CHAPTER 1

Magnetic Circuits

Introduction :

Magnetic flux lines always form closed loops. The closed path followed by the flux lines is called a magnetic circuit. Thus, a magnetic circuit provides a path for magnetic flux, just as an electric circuit provides a path for the flow of electric current. In general, the term magnetic circuit applies to any closed path in space, but in the analysis of electro-mechanical and electronic system this term is specifically used for circuits containing a major portion of ferromagnetic materials. The study of magnetic circuit concepts is essential in the design, analysis and application of electromagnetic devices like transformers, rotating machines, electromagnetic relays etc.

Magnetomotive Force (M.M.F) :

Flux is produced round any current – carrying coil. In order to produce the required flux density, the coil should have the correct number of turns. The product of the current and the number of turns is defined as the coil magneto motive force (m.m.f).

If I = Current through the coil (A)
N = Number of turns in the coil.
Magnetomotive force = Current x turns

So M.M.F = I X N
The unit of M.M.F. is ampere–turn (AT) but it is taken as Ampere(A) since N has no dimensions.

Magnetic Field Intensity:

Magnetic Field Intensity is defined as the magneto-motive force per unit length of the magnetic flux path. Its symbol is H.

$$\text{Magnetic field Intensity (H)} = \frac{\text{Magnetomotive force}}{\text{Mean length of the magnetic path}}$$

- $H = \frac{F}{l} = \frac{I.N.}{l}$ A/m

Where *l* is the mean length of the magnetic circuit in meters. Magnetic field intensity is also called magnetic field strength or magnetizing force.

Permeability :-

Every substance possesses a certain power of conducting magnetic lines of force. For example, iron is better conductor for magnetic lines of force than air (vaccum) . Permeability of a material (μ) is its conducting power for magnetic lines of force. It is the ratio of the flux density. (B) Produced in a material to the magnetic filed strength (H).

$$\text{i.e. } \mu = B/H$$

Reluctance :

Reluctance (s) is akin to resistance (which limits the electric Current).

Flux in a magnetic circuit is limited by reluctance. Thus reluctance(s) is a measure of the opposition offered by a magnetic circuit to the setting up of the flux.

Reluctance is the ratio of magneto motive force to the flux. Thus

$$S = Mmf/\phi$$

Its unit is ampere turns per webber (or AT/wb).

Permeance:-

The reciprocal of reluctance is called the permeance (symbol A).

Permeance (A) = 1/S wb/AT

Turn T has no unit.

Hence permeance is expressed in wb/A or Henerys(H).

B.H. Curve :

Place a piece of an unmagnetised iron bar AB within the field of a solenoid to magnetise it. The field H produced by the solenoid, is called magnetising field, whose value can be altered (increased or decreased) by changing (increasing or decreasing) the current through the solenoid. If we increase slowly the value of magnetic field (H) from zero to maximum value, the value of flux density (B) varies along 1 to 2 as shown in the figure and the magnetic materials (i.e iron bar) finally attains the maximum value of flux density (Bm) at point 2 and thus becomes magnetically saturated.

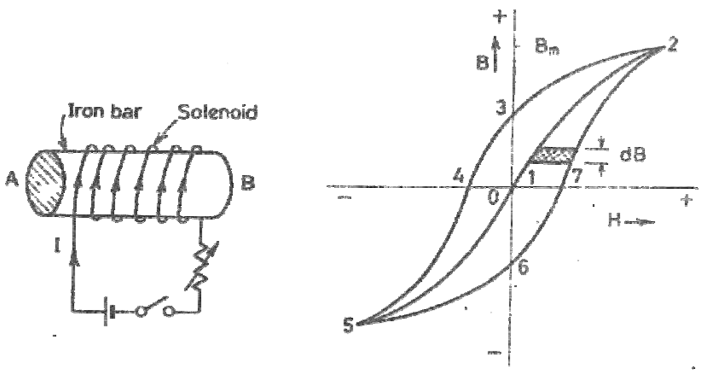


Fig. 2.1

Now if value of H is decreased slowly (by decreasing the current in the solenoid) the corresponding value of flux density (B) does not decreases along 2-1 but decreases some what less rapidly along 2 to 3. Consequently during the reversal of magnetization, the value of B is not zero, but is '13' at H= 0. In other wards, during the period of removal of magnetization force (H), the iron bar is not completely demagnetized.

In order to demagnetise the iron bar completely, we have to supply the demagnetisastion force (H) in the opposite direction (i.e. by reserving the direction of current in the solenoid). The value of B is reduced to zero at point 4, when H='14'. This value of H required to clear off the residual magnetisation, is known as coercive force i.e. the tenacity with which the material holds to its magnetism.

If after obtaining zero value of magnetism, the value of H is made more negative, the iron bar again reaches, finally a state of magnetic saturation at the point 5, which represents negative saturation. Now if the value of H is increased from negative saturation (= '45') to positive saturation (= '12') a curve '5,6,7,2' is obtained. The closed loop "2,3,4,5,6,7,2" thus represents one complete cycle of magnetisation and is known as hysteresis loop.

CHAPTER 02

COUPLED CIRCUITS

It is defined as the interconnected loops of an electric network through the magnetic circuit.

There are two types of induced emf.
(1) Statically Induced emf.
(2) Dynamically Induced emf.

Faraday's Laws of Electro-Magnetic :
Introduction → First Law :→
Whenever the magnetic flux linked with a circuit changes, an emf is induced in it.
OR
Whenever a conductor cuts magnetic flux an emf is induced in it.
Second Law :→
It states that the magnitude of induced emf is equal to the rate of change of flux linkages.
OR
The emf induced is directly proportional to the rate of change of flux and number of turns

Mathematically :

$$e \propto \frac{d\phi}{dt}$$

$$e \propto N$$

Or

$$e = -N\frac{d\phi}{dt}$$

Where e = induced emf
N = No. of turns
ϕ = flux

'- ve' sign is due to Lenz's Law

Inductance :→
It is defined as the property of the substance which opposes any change in Current & flux.

Unit :→ Henry

Fleming's Right Hand Rule:→

It states that "hold your right hand with fore-finger, middle finger and thumb at right angles to each other. If the fore-finger represents the direction of field, thumb represents the direction of motion of the conductor, then the middle finger represents the direction of induced emf."

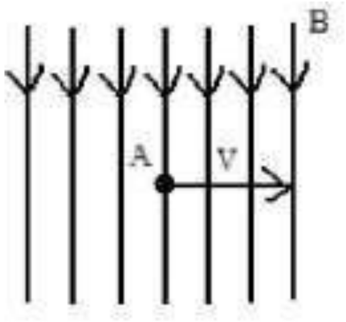
Lenz's Law : →

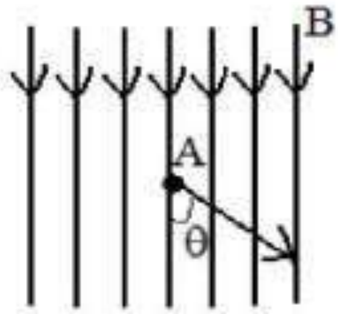
It states that electromagnetically induced current always flows in such a direction that the action of magnetic field set up by it tends to oppose the vary cause which produces it.

OR

It states that the direction of the induced current (emf) is such that it opposes the change of magnetic flux.

(2) **Dynamically Induced emf :→**





In this case the field is stationary and the conductors are rotating in an uniform magnetic field at flux density 'B" Wb/mt2 and the conductor is lying perpendicular to the magnetic field. Let '*l*' is the length of the conductor and it moves a distance of 'dx' nt in time 'dt' second.

The area swept by the conductor = *l. dx*
Hence the flux cut = *ldx. B*
Change in flux in time 'dt' second =

$$\frac{Bldx}{dt}$$

$$E = \mathbf{Blv}$$

If the conductor is making an angle 'θ' with the magnetic field, then
$e = Blv \sin\theta$

(1) Statically Induced emf :→

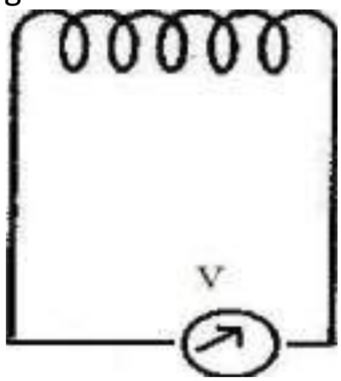
Here the conductors are remain in stationary and flux linked with it changes by increasing or decreasing.

It is divided into two types .

(i) Self-induced emf.

(ii) Mutually-induced emf.

(i) Self-induced emf : → It is defined as the emf induced in a coil due to the change of its own flux linked with the coil.



If current through the coil is changed then the flux linked with its own turn will also change which will produce an emf is called self-induced emf.

Self-Inductance :→

It is defined as the property of the coil due to which it opposes any change (increase or decrease) of current or flux through it.

Co-efficient of Self-Inductance (L) :→

It is defined as the ratio of weber turns per ampere of current in the coil.

OR

It is the ratio of flux linked per ampere of current in the coil.

1st Method for 'L' :→

$$L = \frac{N\phi}{I}$$

Where L = Co-efficient of self-induction

N = Number of turns

ϕ = flux

I = Current

2nd Method for L :→

We know that

$$L = \frac{N\phi}{I}$$

$$\Rightarrow LI = N\phi$$

$$\Rightarrow -LI = -N\phi$$

$$\Rightarrow -L\frac{dI}{dt} = -N\frac{d\phi}{dt}$$

$$\Rightarrow -L\frac{dI}{dt} = -N\frac{d\phi}{dt}$$

$$\Rightarrow -L\frac{dI}{dt} = e_L$$

$$\Rightarrow L\frac{dI}{dt} = -e_L$$

$$\boxed{\Rightarrow L = \frac{-e_L}{\frac{dI}{dt}}}$$

Where L = Inductance

$e_L = -N\frac{d\phi}{dt}$ is known as self-induced emf.

When $\frac{dI}{dt} = 1amp/\sec.$

e = 1 volt

L = 1 Henry

A coil is said to be a self-inductance of 1 Henry if 1 volt is induced in it. When the current through it changes at the rate of 1 amp/ sec.

3rd Method for L :→

$$L = \frac{M_o M_r A N^2}{l}$$

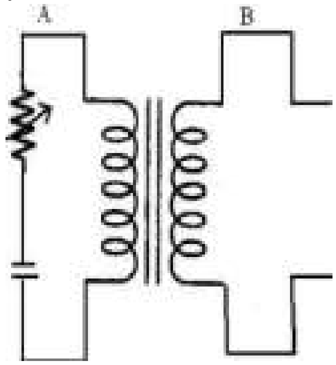
Where A = Area of x-section of the coil

N = Number of turns

L = Length of the coil

(ii) Mutually Induced emf :→

It is defined as the emf induced in one coil due to change in current in other coil. Consider two coils ‘A’ and ‘B’ lying close to each other. An emf will be induced in coil ‘B’ due to change of current in coil ‘A’ by changing the position of the rheostat.



Mutual Inductance :→

It is defined as the emf induced in coil ‘B’ due to change of current in coil ‘A’ is the ratio of flux linkage in coil ‘B’ to 1 amp. Of current in coil ‘A’.

Co-efficient of Mutual Inductance (M):

Coefficient of mutual inductance between the two coils is defined as the weber-turns in one coil due to one ampere current in the other.

1st Method for ‘M’ :→

$$M = \frac{N_2\phi_1}{I_1}$$

N_2 = Number of turns
M = Mutual Inductance
ϕ_1 = flux linkage
I_1 = Current in ampere

2nd Method for M :→

We know that

$$M = \frac{N_2\phi_1}{I_1}$$

$\Rightarrow MI_2 = N_2\,\phi_1$
$\Rightarrow MI_1 = N_2\,\phi_1$

$$\Rightarrow -M\frac{dI_1}{dt} = -N_2\frac{d\phi_1}{dt}$$

$$\Rightarrow -M\frac{dI_1}{dt} = e_M$$

$$\Rightarrow M\frac{dI_1}{dt} = -e_M$$

$$\boxed{\Rightarrow M = \frac{-e_M}{\frac{dI_1}{dt}}}$$

Where

$$e_M = -N_2\frac{d\phi_1}{dt}$$

$e_M = -1$ *VOLT*

Then M = 1 Henry

A coil is said to be a mutual inductance of 1 Henry when 1 volt is induced when the current of 1 amp/sec. is changed in its neighbouring coil.

3rd Method for M :→

$$M = \frac{M_o M_r A N_1 N_2}{l}$$

Co-efficient of Coupling :

Consider two magnetically coupled coils having N_1 and N_2 turns respectively. Their individual co-efficient of self-inductances are

$$L_1 = \frac{M_o M_r A N_1^2}{l}$$

$$L_2 = \frac{M_o M_r A N_2^2}{l}$$

The flux φ_1 produced in coil 'A' due to a current of I_1 ampere is

$$\phi_1 = \frac{L_1 I_1}{N_1} = \frac{M_o M_r A N_1^2}{l} \times \frac{I_1}{N_1}$$

$$\phi_1 = \frac{M_o M_r A N_1 I_1}{l}$$

Suppose a fraction of this flux i.e. $K_1\varphi_1$ is linked with coil 'B'

Then $M = \frac{K_1\phi_1}{I_1} \times N_2 = \frac{K_1 N_1 N_2}{l / M_o M_r A}$ ---------------------------(1)

Similarly the flux φ_2 produced in coil 'B' due to I_2 amp. Is

$$\phi_2 = \frac{M_1 M_r A N_2 I_2}{l}$$

Suppose a fraction of this flux i.e. $K_2\varphi_2$ is linked with coil 'A'

Then $M = \frac{K_2\phi_2}{I_2} \times N_1 = \frac{K_2 N_{21} N_1}{l / M_o M_r A}$ --------------------------(2)

Multiplying equation (1) & (2)

$$M^2 = \frac{K_1 K_2 N_1^2 N_2^2}{l^2 / M_o^2 M_r^2 A^2} \times N_1$$

$$= K^2 \left(\frac{M_o M_r A N_1^2}{l}\right)\left(\frac{M_o M_r A N_2^2}{l}\right)$$

$$[\because K_1 = K_2 = K]$$

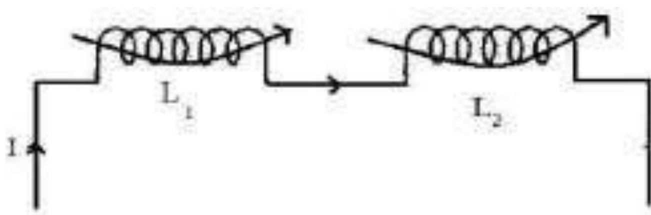
$$M^2 = K^2 . L_1 L_2$$

$$K^2 = \frac{M^2.}{L_1 . L_2}$$

$$\boxed{\Rightarrow K = \sqrt{\frac{M.}{L_1 . L_2}}}$$

Where 'K' is known as the co-efficient of coupling.
Co-efficient of coupling is defined as the ratio of mutual inductance between two coils to the square root of their self- inductances.

Inductances In Series (Additive) :→



oil.

Fluxes are in the same durection

L_2 – Co-efficient of self-inductance of second coil.

EMF induced in first coil due to self-inductance

$$e_{L_1} = -L_1 \frac{dI}{dt}$$

Mutually induced emf in first coil

$$e_{M_1} = -M\frac{dI}{dt}$$

EMF induced in second coil due to self induction

$$e_{L_2} = -L_2\frac{dI}{dt}$$

Mutually induced emf in second coil

$$e_{M_2} = -M\frac{dI}{dt}$$

Total induced emf

$E = e_{L1} + e_{L2} + e_{m1} + e_{m2}$

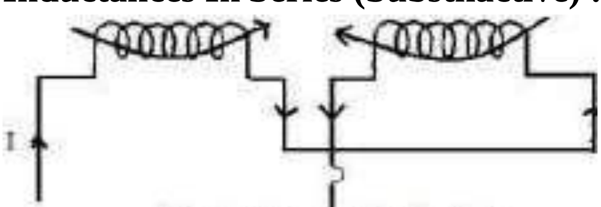
If 'L' is the equivalent inductance, then

$$-L\frac{dI}{dt} = -L_1\frac{dI}{dt} - M\frac{dI}{dt} - L_2\frac{dI}{dt} - M\frac{dI}{dt}$$

$$\Rightarrow -L\frac{dI}{dt} = -\frac{dI}{dt}(L_1 - L_2 - 2M)$$

$$\boxed{\Rightarrow L = L_1 + L_2 + 2M}$$

Inductances In Series (Substnactive) :→



(Fluxes are opposite in direction)

Let M = Co-efficient of mutual inductance

L_1 = Co-efficient of self-inductance of first coil

L_2 -= Co-efficient of self-inductance of second coil

Emf induced in first coil due to self induction

$$e_{L_1} = -L_1\frac{dI}{dt}$$

Mutually induced emf in first coil

$$e_{M_1} = -\left(-M\frac{dI}{dt}\right) = M\frac{dI}{dt}$$

Emf induced in second coil due to self-induction

$$e_{L_2} = -L_2 \frac{dI}{dt}$$

Mutually induced emf in second coil

$$e_{M_2} = -\left(-M \frac{dI}{dt}\right) = M \frac{dI}{dt}$$

Total induced emf

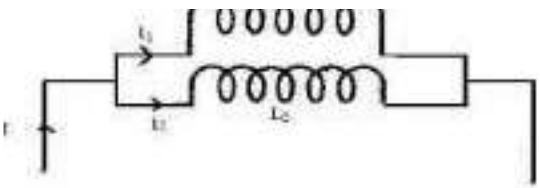
$e = e_{L_1} + e_{L_2} + e_{M_1} + e_{M_2}$

Then

$$-L\frac{dI}{dt} = -L_1\frac{dI}{dt} - L_2\frac{dI}{dt} + M\frac{dI}{dt} + M\frac{dI}{dt}$$

$$\Rightarrow -L\frac{dI}{dt} = -\frac{dI}{dt}(L_1 + L_2 - 2M) \qquad \Rightarrow L = L_1 + L_2 - 2M$$

Inductances In Parallel :→



Let two inductances of L_1 & L_2 are connected in parallel

Let the co-efficent of mutual inductance between them is M.

$I-i_1+i_2$

$$\frac{dI}{dt} = \frac{di_1}{dt} + \frac{di_2}{dt} \text{ ------------------------------ (1)}$$

$$e = L_1\frac{di_1}{dt} + M\frac{di_2}{dt}$$

$$= L_2\frac{di_2}{dt} + M\frac{di_1}{dt}$$

$$\Rightarrow L_1\frac{di_1}{dt} + M\frac{di_2}{dt} = L_2\frac{di_2}{dt} + M\frac{di_1}{dt}$$

$$\Rightarrow (L_1 - M)\frac{di_1}{dt} = (L_2 - M)\frac{di_2}{dt}$$

$$\Rightarrow \frac{di_1}{dt} = \frac{(L_2 - M)}{(L_1 - M)}\frac{di_2}{dt} \text{ --------------------------------(2)}$$

$$\frac{dI}{dt} = \frac{di_1}{dt} + \frac{di_2}{dt}$$

$$= \frac{(L_2 - M)}{(L_1 - M)}\frac{di_2}{dt} + \frac{di_2}{dt}$$

$$\Rightarrow \frac{dI}{dt} = \left(\frac{L_2 - M}{L_1 - M} + 1\right)\frac{di_2}{dt} \text{ ---------------------------(3)}$$

If 'L' is the equivalent inductance

$$e = L\frac{di}{dt} = L_1\frac{di_1}{dt} + M\frac{di_2}{dt}$$

$$L\frac{di}{dt} = L_1\frac{di_1}{dt} + M\frac{di_2}{dt}$$

$$\Rightarrow \frac{di}{dt} = \frac{1}{L}\left(L_1\frac{di_1}{dt} + M\frac{di_2}{dt}\right) \text{------------------------}(4)$$

Substituting the value of $\frac{di_1}{dt}$

$$\frac{di}{dt} = \frac{1}{L}\left[L_1\frac{L_2 - M}{L_1 - M} + M\right]\frac{di_2}{dt} \text{------------------------}(5)$$

Equating equation (3) & (5)

$$\left[\left(\frac{L_2 - M}{L_1 - M}\right) + 1\right]\frac{di_2}{dt} = \frac{1}{L}\left[L_1\left(\frac{L_2 - M}{L_1 - M}\right) + M\right]\frac{di_2}{dt}$$

$$\Rightarrow \frac{L_2 - M}{L_1 - M} + 1 = \frac{1}{L}\left[L_1\left(\frac{L_2 - M}{L_1 - M}\right) + M\right]$$

$$\Rightarrow \frac{L_2 - M + L_1 - M}{L_1 - M} = \frac{1}{L}\left[\frac{L_1L_2 - L_1M + L_1M - M^2}{L_1 - M}\right]$$

$$\Rightarrow \frac{L_1 + L_2 - 2M}{L_1 - M} = \frac{1}{L}\left[\frac{L_1L_2 - M^2}{L_1 - M}\right]$$

$$\Rightarrow L_1 + L_2 - 2M = \frac{1}{L}\left[L_1L_2 - M^2\right]$$

$$\boxed{\Rightarrow L = \frac{L_1L_2 - M^2}{L_1 + L_2 - 2M}}$$

When mutual field assist.

$$L = \frac{L_1L_2 - M^2}{L_1 + L_2 + 2M}$$

When mutual field opposes.

Exp. -01 :

Two coupled cols have self inductances L_1= 10×10^{-3}H and L_2= 20×10^{-3}H. The coefficient of coupling (K) being 0.75 in the air, find voltage in the second coil and the flux of first coil provided the second coils has 500 turns and the circuit current is given by i_1 = 2sin 314.1A.

Solution :

$M=K\sqrt{L1L2}$

$M = 0.75\ \sqrt{10\times10^{-3}\times 20\times10^{-3}}$

$\Rightarrow M = 10.6\times10^{-3}H$

The voltage induced in second coil is

$V_2 = M\frac{di1}{dt}$

$=10.6\times10^{-3}\times 2\times314\ \cos\ 314dt.$

The magnetic circuit being linear,

$$M = \frac{N_2\phi_2}{i_1} = \frac{500\times(K\phi_1)}{i_1}$$

$$\phi = \frac{M}{500\times K}\times i_1 = \frac{10.6\times10^{-3}}{500\times0.75}\times 2\sin 314t$$

$$= 5.66\times10^{-5}\sin 314t$$

Exp. 02

Find the total inductance of the three series connected coupled coils.Where the self and mutual inductances are

L_1 = 1H, L_2 = 2H, L_3 = 5H

M_{12}= 0.5H, M_{23} = 1H, M_{13} = 1H

Solution:

$L_A = L_1 + M_{12} + M_{13}$

= 1 + 20.5 +1

= 2.5H

$L_B = L_2 + M_{23} + M_{12}$

= 2 + 1 + 0.5

= 3.5H

$L_C = L_3 + M_{23} + M_{13}$

= 5 + 1 + 1

= 7H

Total inductances are

$L_{ea} = L_A + L_B + L_C$

$= 2.5 + 3.5 + 7$

$= 13H$ (Ans)

CHAPTER 3

Circuit Elements and analysis

1.1 Voltage

Energy is required for the movement of charge from one point to another. Let W Joules of energy be required to move positive charge Q columbs from a point a to point b in a circuit. We say that a voltage exists between the two points. The voltage V between two points may be defined in terms of energy that would be required if a charge were transferred from one point to the other. Thus, there can be a voltage between two points even if no charge is actually moving from one to the other. Voltage between a and b is given by

$$V = \frac{W}{Q} J/C$$

$$\text{Hence Electric Potential (V)} = \frac{\text{Worked are (W) in Joules}}{\text{Ch arge (Q)in columbs}}$$

Current :

An electric current is the movement of electric charges along a definite path. In case of a conductor the moving charges are electrons.

The unit of current is the ampere. The ampere is defined as that current which when flowing in two infinitely long parallel conductors of negligible cross section, situated 1 meter apart in Vacuum, produces between the conductors a force of 2 x 10^{-7} Newton per metre length.

Power : Power is defined as the work done per unit time. If a field F newton acts for t seconds through a distance d metres along a straight line, work done W = Fxd N.m. or J. The power P, either generated or dissipated by the circuit element.

$$P = \frac{w}{t} = \frac{F \text{ x } d}{t}$$

Power can also be written as Power = $\frac{\text{Work}}{\text{time}}$

$$= \frac{\text{Work}}{\text{Charge}} \times \frac{\text{Charge}}{\text{Time}} = \text{Voltage x Current}$$

P = V x I watt.

Energy : Electric energy W is defined as the Power Consumed in a given time. Hence, if current IA flows in an element over a time period t second, when a voltage V volts is applied across it, the energy consumed is given by

W = P x t = V x I x t J or watt. second.

The unit of energy W is Joule (J) or watt. second. However, in practice, the unit of energy is kilowatt. hour (Kwh)

1.2 Resistance : According to Ohm's law potential difference (V) across the ends of a conductor is proportional to the current (I) flowing through the conductor at a constant temperature. Mathematically Ohm's law is expressed as

$V \alpha I$ or $V = R \times I$

Or $R = \frac{V}{I}$ where R is the proportionality constant and is designated as the conductor resistance and has the unit of Ohm (Ω).

Conductance : Voltage is induced in a stationary conductor when placed in a varying magnetic field. The induced voltage (e) is proportional to the time rate of change of current, di/dt producing the magnetic field.

Therefore $e \; \alpha \; \frac{di}{dt}$

Or $e = L \frac{di}{dt}$

e and i are both function of time. The proportionality constant L is called inductance. The Unit of inductance is Henery (H).

Capacitance : A capacitor is a Physical device, which when polarized by an electric field by applying a suitable voltage across it, stores energy in the form of a charge separation.

The ability of the capacitor to store charge is measured in terms of capacitance. Capacitence of a capacitor is defined as the charge stored per Volt applied.

$$C = \frac{q}{v} = \frac{\text{Coulomb}}{\text{Volt}} = \text{Farad}$$

1.3 Active and passive Branch :

A branch is said to be active when it contains one or more energy sources. A passive branch does not contain an energy source.

Branch : A branch is an element of the network having only two terminals.

Bilateral and unilateral element :

A bilateral element conducts equally well in either direction. Resistors and inductors are examples of bilateral elements. When the current voltage relations are different for the two directions of current flow, the element is said to be unilateral. Diode is an unilateral element.

Linear Elements : When the current and voltage relationship in an element can be simulated by a linear equation either algebraic, differential or integral type, the element is said to be linear element.

Non Linear Elements : When the current and voltage relationship in an element can not be simulated by a linear equation, the element is said to be non linear elements.

1.4 Kirchhoff's Voltage Law (KVL) :

The algebraic sum of Voltages (or voltage drops) in any closed path or loop is Zero.

Application of KVL with series connected voltage source.

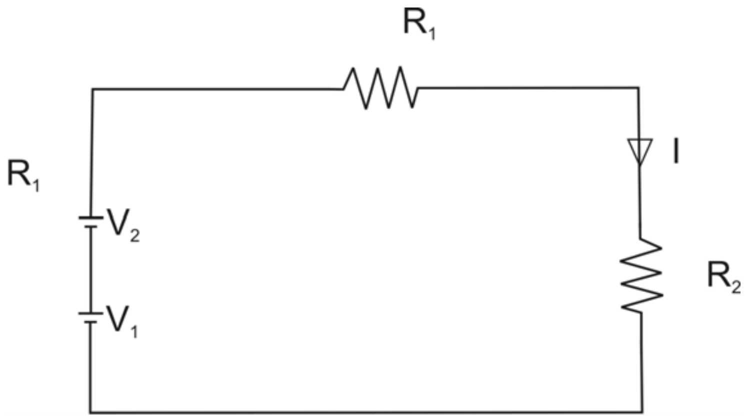


Fig. 1.1

$V_1 + V_2 – IR_1 – IR_2 = 0$

$= V_1 + V_2 = I\ (R_1 + R_2)$

$$I = \frac{V_1 + V_2}{R_1 + R_2}$$

Application of KVL while voltage sources are connected in opposite polarity.

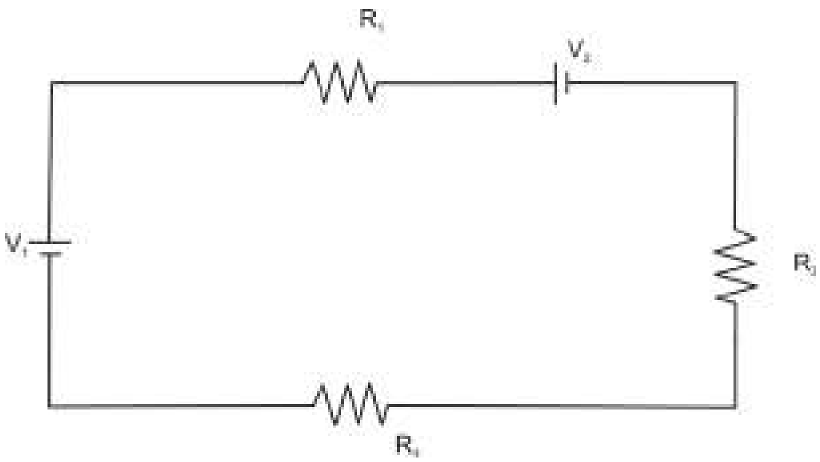


Fig. 1.2

$V_1 – IR_1 – V_2 – IR_2 – IR_3 = 0$

- $V_1 – V_2 = IR_1 + IR_2 + IR_3$
- $V_1 – V_2 = I\ (R_1 + IR_2 + IR_3)$

➢ $$I = \frac{V_1 - V_2}{R_1 + R_2 + R_3}$$

Kirchaoff's Current Law (KCL) :

The algebraic sum of currents meeting at a junction or mode is zero.

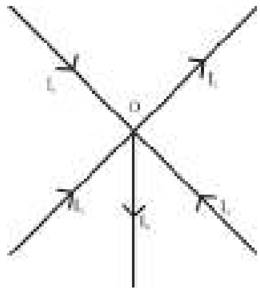


Fig. 1.3

Considering five conductors, carrying currents I_1, I_2, I_3, I_4 and I_5 meeting at a point O. Assuming the incoming currents to be positive and outgoing currents negative.

$$I_1 + (-I_2) + I_3 + (-I_4) + I_5 = 0$$

$$I_1 - I_2 + I_3 - I_4 + I_5 = 0$$

$$I_1 + I_3 + I_5 = I_2 + I_4$$

Thus above Law can also be stated as the sum of currents flowing towards any junction in an electric circuit is equal to the sum of the currents flowing away from that junction.

Voltage Division (Series Circuit)

Considering a voltage source (E) with resistors R_1 and R_2 in series across it.

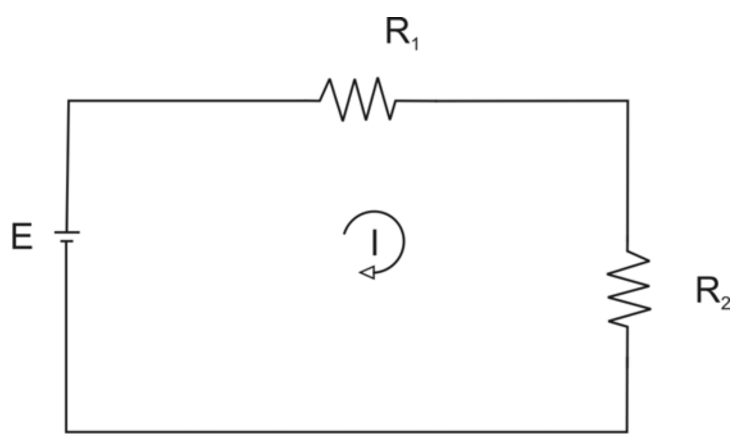


Fig. 1.4

$$I = \frac{E\ R_1}{+ R_2}$$

Voltage drop across R_1 = I. R_1 = $\frac{E.R_1}{R_1 + R_2}$

Similarly voltage drop across R_2 = I.R_2 = $\frac{E.R_1}{R_1 + R_2}$

Current Division :

A parallel circuit acts as a current divider as the current divides in all branches in a parallel circuit.

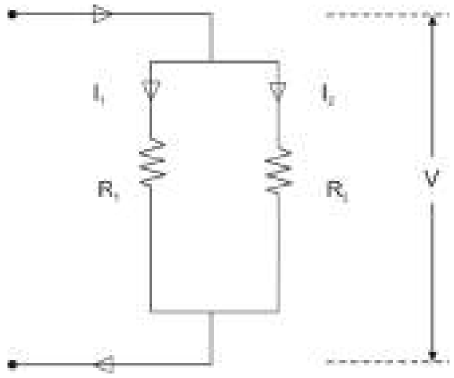


Fig. 1.5

Fig. shown the current I has been divided into I_1 and I_2 in two parallel branches with resistances R_1 and R_2 while V is the voltage drop across R_1 and R_2.

$$I_1 = \frac{V}{R_1} \text{ and } I_2 = \frac{V}{R_2}$$

Let R = Total resistance of the circuit.

Hence $\frac{1}{R} = = \frac{1}{R_1} + \frac{1}{R_2}$

➢ $R = \frac{R_1 R_2}{R_1 + R_2}$

$$I = \frac{V}{R} = \frac{V}{\frac{R_1 R_2}{R_1 + R_2}} = \frac{V(R_1 + R_2)}{R_1 R_2}$$

But = V = $I_1R_1 = I_2R_2$

- $I = I_1R_1\left(\frac{R_1R_2}{R_1 + R_2}\right)$
- $I = \frac{I_1(R_1 + R_2)}{R_2}$

Therefore $\boxed{I_1 = \frac{IR_2}{R_1 + R_2}}$

Similarly it can be derived that

$$\boxed{I_2 = \frac{IR_1}{R_1 + R_2}}$$

NETWORK ANALYSIS

Different terms are defined below:

1. **Circuit:** A circuit is a closed conducting path through which an electric current either . flow or is intended flow

2. **Network:** A combination of various electric elements, connected in any manner. Whatsoever, is called an electric network

3. **Node:** it is an equipotential point at which two or more circuit elements are joined.

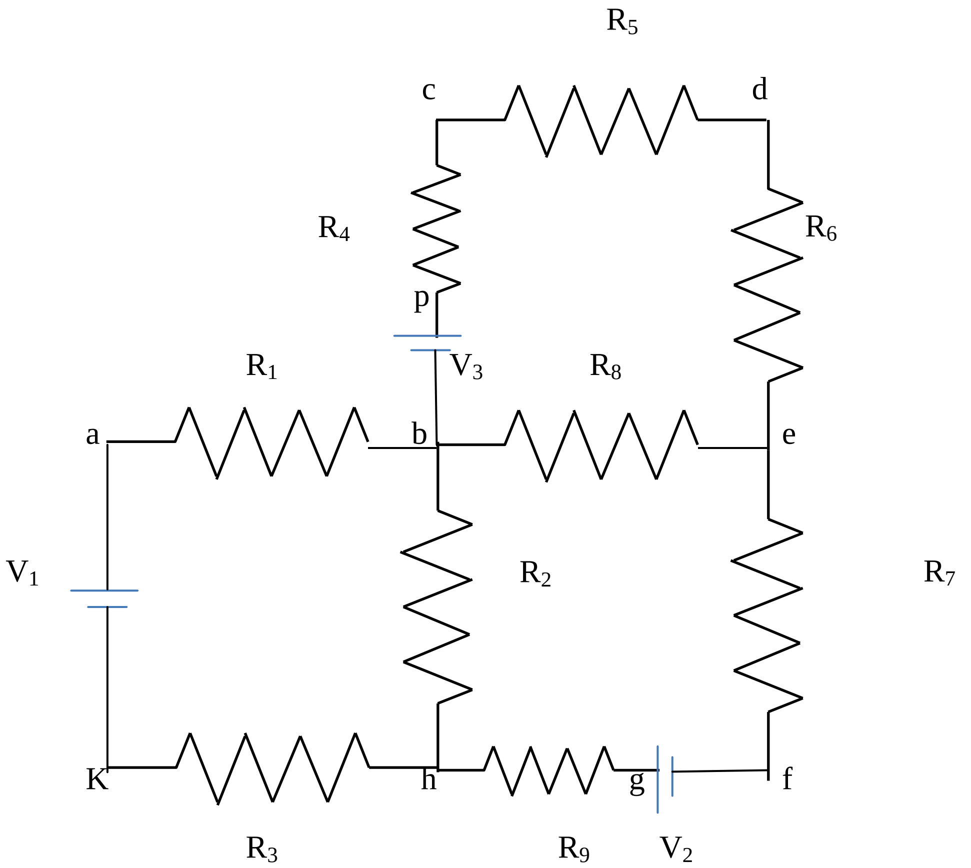
4. **Junction:** it is that point of a network where three or more circuit elements are joined.

5. **Branch:** it is a part of a network which lies between junction points.

6. **Loop:** It is a closed path in a circuit in which no element or node is accounted more than once.

7. **Mesh:** It is a loop that contains no other loop within it.

Example 3.1 In this circuit configuration of figure 3.1, obtain the no. of i) circuit elements ii) nodes iii) junction points iv) branches and v) meshes.



Solution: i) no. of circuit elements = 12 (9 resistors + 3 voltage sources)

ii) no. of nodes =10 (a, b, c, d, e, f, g, h, k, p)

iii) no. of junction points =3 (b, e, h)

iv) no. of branches = 5 (bcde, be, bh, befgh, bakh)

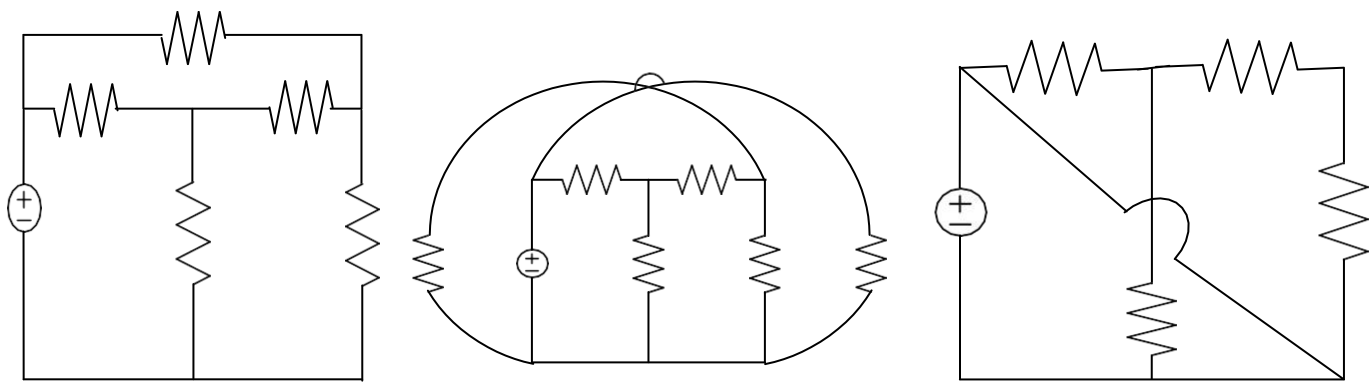
v) no. of meshes = 3 (abhk, bcde, befh)

3.2 MESH ANALYSIS

Mesh and nodal analysis are two basic important techniques used in finding solutions for a network. The suitability of either mesh or nodal analysis to a particular problem depends mainly on the number of voltage sources or current sources .If a network has a large number of voltage sources, it is useful to use mesh analysis; as this analysis requires that all the sources in a circuit be voltage sources. Therefore, if there are any current sources in a circuit they are to be converted into equivalent voltage sources,if, on the other hand, the network has more current sources,nodal analysis is more useful.

Mesh analysis is applicable only for planar networks. For non-planar circuits mesh analysis is not applicable .A circuit is said to be planar, if it can be drawn on a plane surface without crossovers. A non-planar circuit cannot be drawn on a plane surface without a crossover.

Figure 3.2 (a) is a planar circuit. Figure 3.2 (b) is a non-planar circuit and fig. 3.2 (c) is a planar circuit which looks like a non-planar circuit. It has already been discussed that a loop is a closed path. A mesh is defined as a loop which does not contain any other loops within it. To apply mesh analysis, our first step is to check whether the circuit is planar or not and the second is to select mesh currents. Finally, writing Kirchhoff‛s voltage law equations in terms of unknowns and solving them leads to the final solution.



(a) (b) (c)

Figure 3.2

Observation of the Fig.3.2 indicates that there are two loops abefa,and bcdeb in the network .Let us assume loop currents I_1 and $I_2$with directions as indicated in the figure.

Considering the loop abefa alone, we observe that current I_1 is passing through R_1, and (I_1-I_2) is passing through R_2. By applying Kirchhoff's voltage law, we can write

$$V_s = I_1R_1 + R_2(I_1-I_2) \tag{3.1}$$

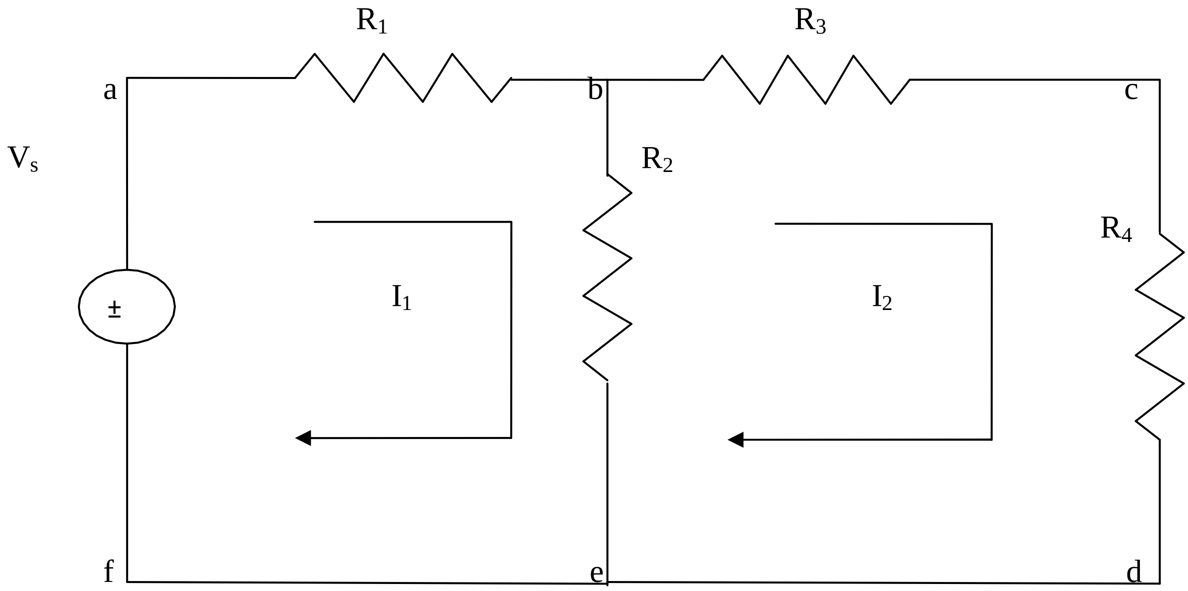


Figure 3.3

Similarly, if we consider the second mesh bcdeb, the current I_2 is passing through R_3 and R_4, and $(I_2 - I_1)$ is passing through R_2. By applying Kirchhoff's voltage law around the second mesh, we have

$$R_2 (I_2-I_1) + R_3I_2 + R_4I_2 = 0 \tag{3.2}$$

By rearranging the above equations,the corresponding mesh current equations are

$$I_1 (R_1+R_2) - I_2R_2 = V_s$$

$$-I_1R_2 + (R_2+R_3+R_4) I_2 = 0 \tag{3.3}$$

By solving the above equations, we can find the currents I_1 and I_2.If we observe Fig.3.3, the circuit consists of five branches and four nodes, including the reference node.The number of mesh currents is equal to the number of mesh equations.

And the number of equations=branches-(nodes-1).in Fig.3.3, the required number of mesh current would be 5-(4-1)=2.

In general we have B number of branches and N number of nodes including the reference node than number of linearly independent mesh equations M=B-(N-1).

Example 3.2 Write the mesh current equations in the circuit shown in fig 3.4 and determine the currents.

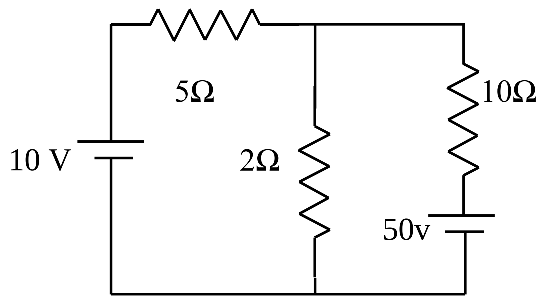


Figure 3.4

Solution: Assume two mesh currents in the direction as indicated in fig. 3.5. The mesh current equations are

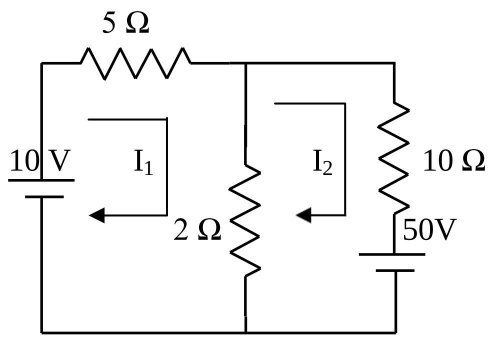


Figure 3.5

$5I_1 + 2(I_1-I_2) = 10$

$101_2 + 2(1_2-1_1) + 50= 0$ (3.4)

We can rearrange the above equations as

$7I_1 -2I_2 =10$

$-2I_1+12I_2 = -50$ (3.5)

By solving the above equations, we have I_1= 0.25 A, and I_2 = -4.125

Here the current in the second mesh $I_{2,}$ is negative; that is the actual current I_2 flows opposite to the assumed direction of current in the circuit of fig .3.5.

Example 3.3 Determine the mesh current I_1 in the circuit shown in fig.3.6.

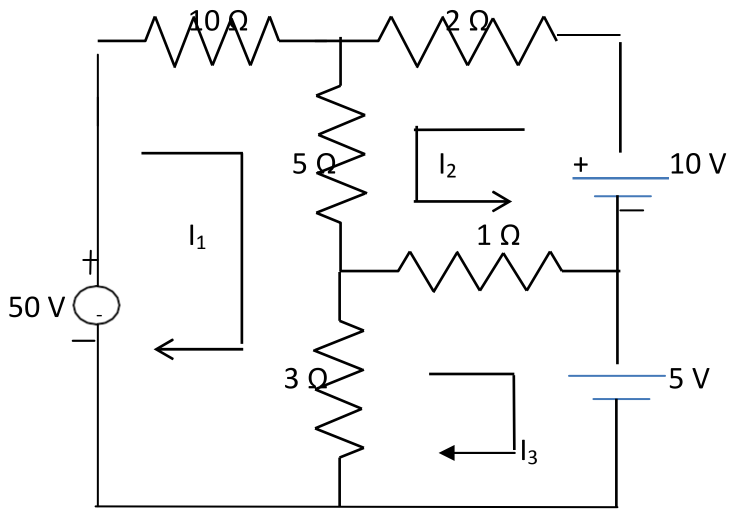


Figure 3.6

Solution: From the circuit, we can from the following three mesh equations

$$10I_1+5(I_1+I_2)+3(I_1-I_3) = 50 \quad (3.6)$$

$$2I_2+5(I_2+I_1)+1(I_2+I_3) = 10 \quad (3.7)$$

$$3(I_3-I_1)+1(I_3+I_2) = -5 \quad (3.8)$$

Rearranging the above equations we get

$$18I_1+5I_2-3I_3=50 \quad (3.9)$$

$$5I_1+8I_2+I_3=10 \quad (3.10)$$

$$-3I_1+I_2+4I_3=-5 \quad (3.11)$$

According to the Cramer's rule

$$I_1 = \frac{\begin{vmatrix} 50 & 5 & -3 \\ 10 & 8 & 1 \\ -5 & 1 & 4 \end{vmatrix}}{\begin{vmatrix} 18 & 5 & -3 \\ 5 & 8 & 1 \\ -3 & 1 & 4 \end{vmatrix}} = \frac{1175}{356}$$

Or I_1= 3.3 A Similarly,

$$I_2 = \frac{\begin{vmatrix} 18 & 50 & -3 \\ 5 & 10 & 1 \\ -3 & -5 & 4 \end{vmatrix}}{\begin{vmatrix} 18 & 5 & -3 \\ 5 & 8 & 1 \\ -3 & 1 & 4 \end{vmatrix}} = \frac{-355}{356}$$

Or I_2=-0.997A (3.12)

$$I_3 = \frac{\begin{vmatrix} 18 & 5 & 50 \\ 5 & 8 & 10 \\ -3 & 1 & -5 \end{vmatrix}}{\begin{vmatrix} 18 & 5 & -3 \\ 5 & 8 & 1 \\ -3 & 1 & 4 \end{vmatrix}} = \frac{525}{356}$$

Or I_3=1.47A (3.13)

$\therefore I_1=3.3A,\ I_2=-0.997A,\ I_3=1.47A$

3.3 MESH EQUATIONS BY INSPECTION METHOD The mesh equations for a general planar network can be written by inspection without going through the detailed steps. Consider a three mesh networks as shown in figure 3.7

The loop equation are $I_1R_1 + R_2(I_1-I_2) = V_1$

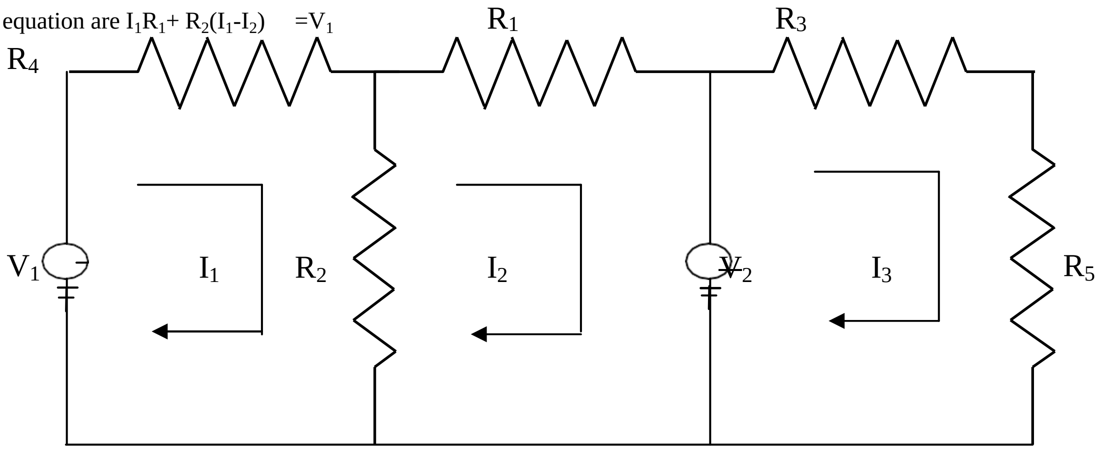


Figure 3.7

$R_2(I_2-I_1)+I_2R_3= -V_2$ 3.14

$R_4I_3+R_5I_3=V_2$ 3.15

Reordering the above equations, we have

$(R_1+R_2)I_1-R_2I_2=V_1$ 3.16

$-R_2I_1+(R_2+R_3)I_2=-V_2$ 3.17

$(R_4+R_5)I_3=V_2$ 3.18

The general mesh equations for three mesh resistive network can be written as

$R_{11}I_1 \pm R_{12}I_2 \pm R_{13}I_3= V_a$ 3.19

$\pm R_{21}I_1+R_{22}I_2 \pm R_{23}I_3= V_b$ 3.20

$\pm R_{31}I_1 \pm R_{32}I_2+R_{33}I_3= V_c$ 3.21

By comparing the equations 3.16, 3.17 and 3.18 with equations 3.19, 3.20 and 3.21 respectively, the following observations can be taken into account.

1. The self-resistance in each mesh
2. The mutual resistances between all pairs of meshes and
3. The algebraic sum of the voltages in each mesh.

The self-resistance of loop 1, $R_{11}=R_1+R_2$, is the sum of the resistances through which I_1 passes.

The mutual resistance of loop 1, $R_{12}= -R_2$, is the sum of the resistances common to loop currents I_1 and I_2. If the directions of the currents passing through the common resistances are the same, the mutual resistance will have a positive sign; and if the directions of the currents passing through the common resistance are opposite then the mutual resistance will have a negative sign.

$V_a=V_1$ is the voltage which drives the loop 1. Here the positive sign is used if the direction of the currents is the same as the direction of the source. If the current direction is opposite to the direction of the source, then the negative sign is used.

Similarly $R_{22}=R_2+R_3$ and $R_{33}=R_4+R_5$ are the self-resistances of loops 2 and 3 respectively. The mutual resistances $R_{13}=0$, $R_{21}= -R_2$, $R_{23}=0$, $R_{31}=0$, $R_{32}=0$ are the sums of the resistances common to the mesh currents indicated in their subscripts.

$V_b= -V_2$, $V_c= V_2$ are the sum of the voltages driving their respective loops.

Example 3.4 write the mesh equation for the circuit shown in fig. 3.8

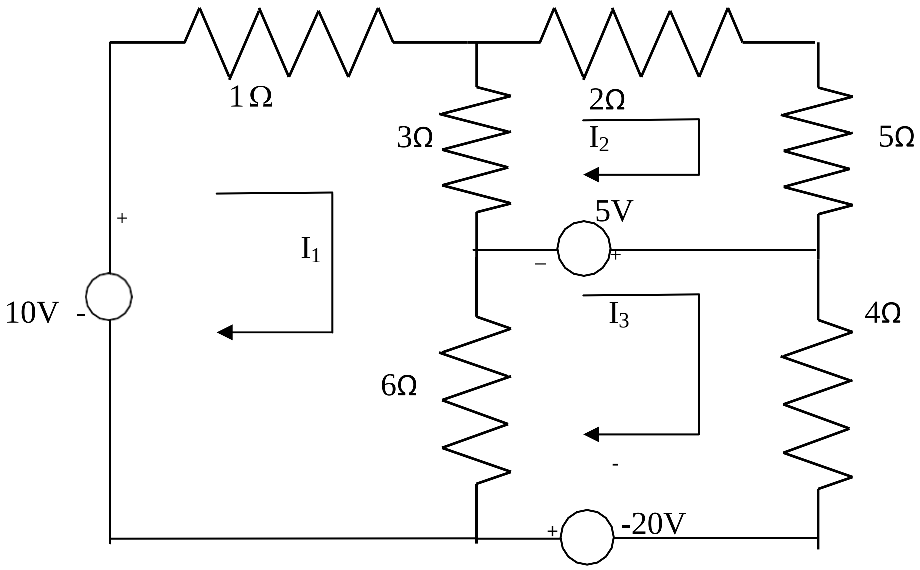


Figure 3.8

Solution : the general equation for three mesh equation are

$$R_{11}I_1 \pm R_{12}I_2 \pm R_{13}I_3 = V_a \quad (3.22)$$

$$\pm R_{21}I_1 + R_{22}I_2 \pm R_{23}I_3 = V_b \quad (3.23)$$

$$\pm R_{31}I_1 \pm R_{32}I_2 + R_{33}I_3 = V_c \quad (3.24)$$

Consider equation 3.22

R_{11}=self resistance of loop 1=(1Ω+ 3 Ω +6 Ω) =10 Ω

R_{12}= the mutual resistance common to loop 1 and loop 2 = -3 Ω

Here the negative sign indicates that the currents are in opposite direction .

R_{13}= the mutual resistance common to loop 1 & 3= -6 Ω

V_a= +10 V, the voltage the driving the loop 1.

Here he positive sign indicates the loop current I_1 is in the same direction as the source element.

Therefore equation 3.22 can be written as

$10\ I_1 - 3I_2 - 6I_3 = 10\ V$ (3.25)

Consider Eq. 3.23

R_{21}= the mutual resistance common to loop 1 and loop 2 = -3 Ω

R_{22}= self resistance of loop 2=(3Ω+ 2 Ω +5 Ω) =10 Ω

R_{23}=0, there is no common resistance between loop 2 and 3.

V_b = -5 V, the voltage driving the loop 2.

Therefore Eq. 3.23 can be written as

$-3I_1 + 10I_2 = -5V$ (3.26)

Consider Eq. 3.24

R_{31}= the mutual resistance common to loop 1 and loop 3 = -6 Ω

R_{32}= the mutual resistance common to loop 3 and loop 2 = 0

R_{33}= self resistance of loop 3=(6Ω+ 4 Ω) =10 Ω

V_c= the algebraic sum of the voltage driving loop 3

=(5 V+20V)=25 V (3.27)

Therefore, Eq3.24can be written as $-6I_1 + 10I_3 = 25V$

$-6I_1 - 3I_2 - 6I_3 = 10V$

$-3I_1 + 10I_2 = -5V$

$-6I_1 + 10I_3 = 25V$

3.4 SUPERMESH ANALYSIS

Suppose any of the branches in the network has a current source, then it is slightly difficult to apply mesh analysis straight forward because first we should assume an unknown voltage across the current source, writing mesh equation as before, and then relate the source current to the assigned mesh currents. This is generally a difficult approach. On way to overcome this difficulty is by applying the supermesh technique. Here we have to choose the kind of supermesh. A supermesh is constituted by two adjacent loops that have a common current source. As an example, consider the network shown in the figure 3.9.

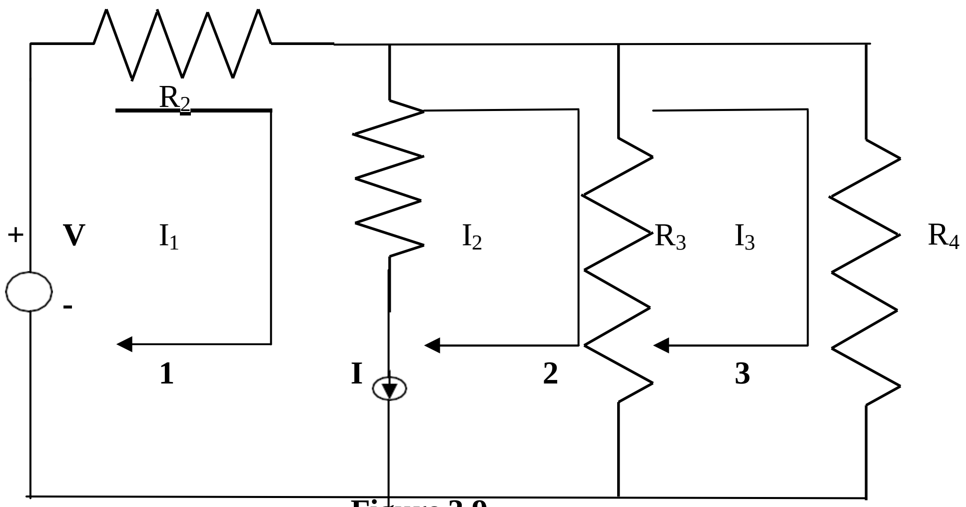


Figure 3.9

Here the current source I is in the common boundary for the two meshes 1 and 2. This current source creates a supermesh, which is nothing but a combination of meshes 1 and 2.

$R_1I_1 + R_3(I_2-I_3)=V$

Or $R_1I_1 + R_3I_2 - R_4I_3= V$

Considering mesh 3, we have

$R_3(I_3-I_2)+ R_4I_3=0$

Finally the current I from current source is equal to the difference between two mesh currents i.e.

$I_1-I_2=I$

we have thus formed three mesh equations which we can solve for the three unknown currents in the network.

Example 3.5. Determine the current in the 5Ω resistor in the network given in Fig. 3.10

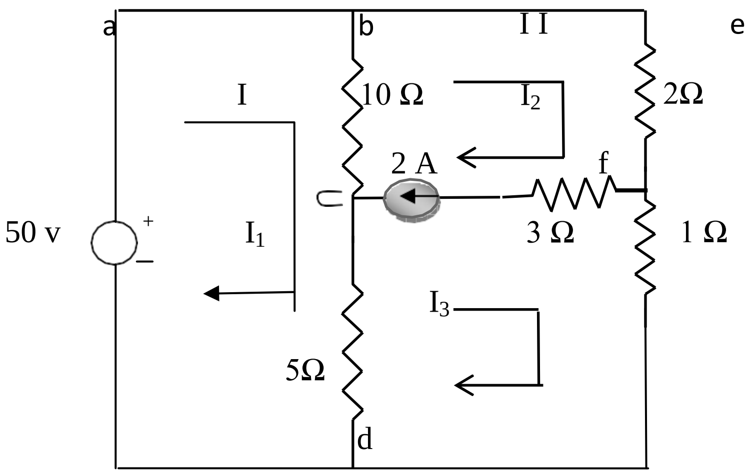


Figure 3.10

Solution: - From the first mesh, i.e. abcda, we have

$$50 = 10(I_1-I_2) + 5(I_1-I_3)$$

$$\text{Or}\quad 15I_1-10I_2 -5I_3 =50 \qquad (3.28)$$

From the second and third meshes. we can form a super mesh

$$10(I_2-I_1)+2I_2 +I_3+5(I_3-I_1)=0$$

$$\text{Or}\quad -15I_1+12I_2 +6I_3 =0 \qquad (3.29)$$

The current source is equal to the difference between II and III mesh currents

i.e. I_2-I_3 = 2A (3.30)

Solving 3.28.,3.29 and 3.30. we have

I_1 =19.99A,I_2= 17.33 A, and I_3 = 15.33 A

The current in the 5Ω resistor =I_1 -I_3

=19.99 -15.33=4.66A

The current in the 5Ω resistor is 4.66A.

Example 3.6. Write the mesh equations for the circuit shown in fig. 3.11 and determine the currents, I_1, I_2 and I_3.

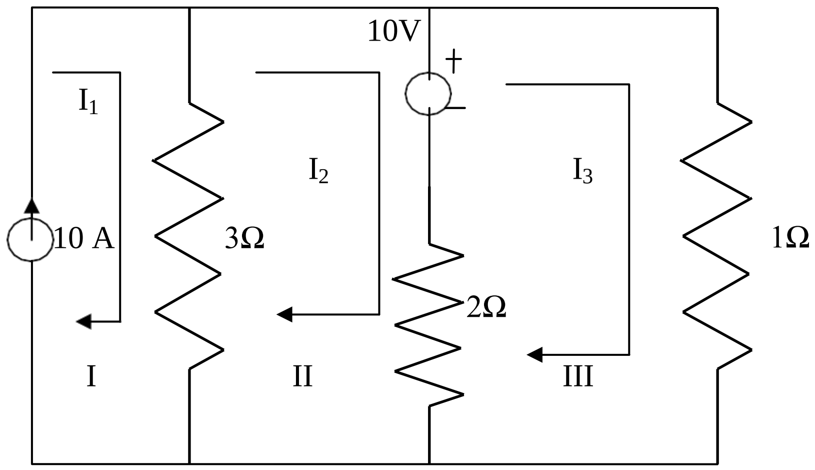


Figure 3.11

Solution ; In fig 3.11, the current source lies on the perimeter of the circuit, and the first mesh is ignored. Kirchhoff's voltage law is applied only for second and third meshes .

From the second mesh, we have

$3(I_2-I_1)+2(I_2-I_3)+10 =0$

Or $-3I_1 +5I_2-2I_3 = -10$ (3.31)

From the third mesh, we have

$I_3 + 2 (I_3 -I_2) =10$

Or $-2I_2+3I_3 =10$ (3.32)

From the first mesh, I_1 =10A (3.33)

From the abovethree equations, we get

I_1=10A, I_2 =7.27, I_3 =8.18A

3.5 NODALANALYSIS

In the chapter I we discussed simple circuits containing only two nodes, including the reference node. In general, in a N node circuit, one of the nodes is chosen as the reference or datum node, then it is possible to write N -1nodal equations by assuming N-1 node voltages. For example,a10 node circuit requires nine unknown voltages and nine equations. Each node in a circuit can be assigned a number or a letter. The node voltage is the voltage of a given node with respect to one particular node, called the reference node, which we assume at zero potential. In the circuit shown in fig. 3.12, node 3 is assumed as the Reference node. The voltage at node 1 is the voltage at that node with respect to node 3. Similarly, the voltage at node 2 is the voltage at that node with respect to node 3. Applying Kirchhoff's current law at node 1, the current entering is the current leaving (See Fig.3.13)

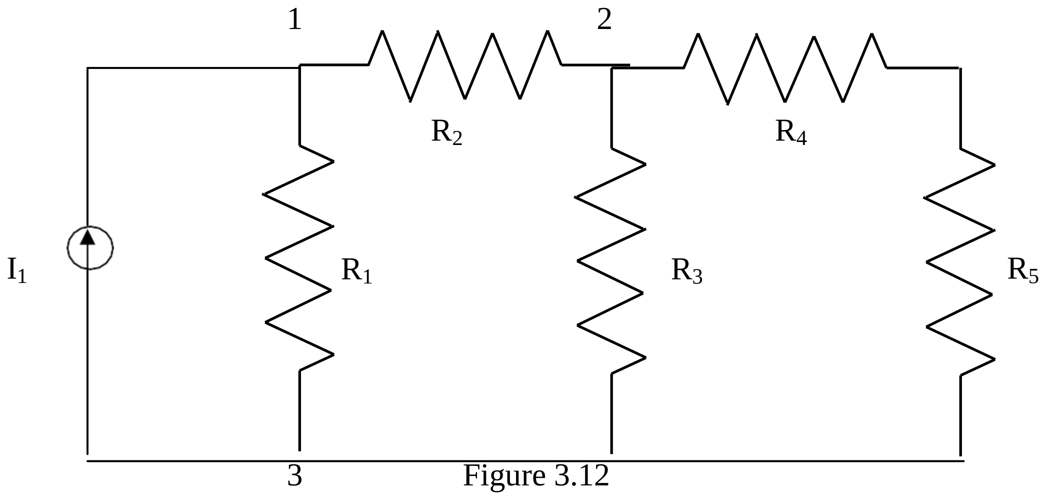


Figure 3.12

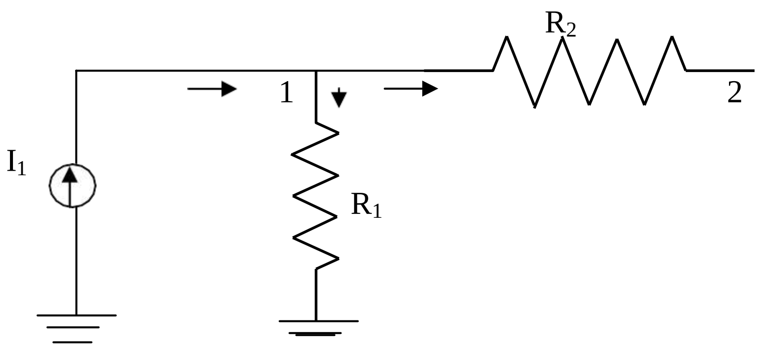


Figure 3.13

$I_1 = V_1/R_1 + (V_1-V_2)/R_2$

Where V_1 and V_2 are the voltages at node 1 and 2, respectively. Similarly, at node 2.the current entering is equal to the current leaving as shown in fig. 3.14

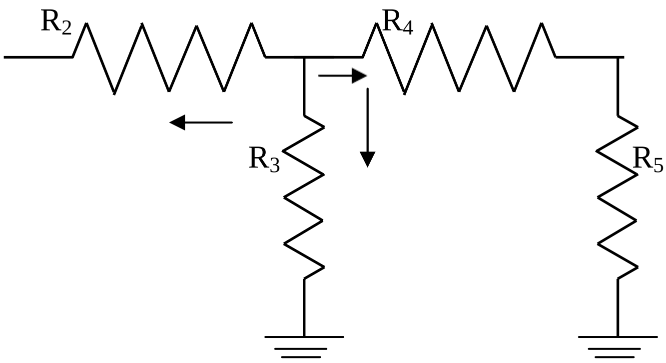


Figure 3.14

$(V_2-V_1)/R_2 + V_2/R_3 + V_2/(R_4+R_5) =0$

Rearranging the above equations, we have

$V_1[1/R_1+1/R_2]-V_2(1/R_2)= I_1$

$-V_1(1/R_2) + V_2[1/R_2+1/R_3+1/(R_4+R_5)]=0$

From the above equations we can find the voltages at each node.

Example 3.7 Determine the voltages at each node for the circuit shown in fig 3.15

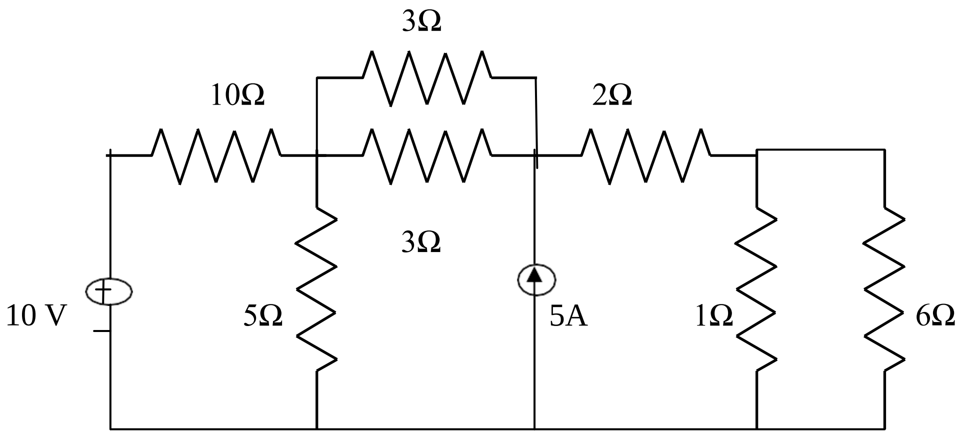


Figure 3.15

Solution : At node 1, assuming that all currents are leaving, we have

$$(V_1-10)/10 + (V_1-V_2)/3 +V_1/5 + (V_1-V_2)/3 =0$$

Or $\quad V_1[1/10 +1/3 +1/5 + 1/3] - V_2[1/3 + 1/3] = 1$

$$0.96V_1-0.66V_2 = 1 \qquad (3.36)$$

At node 2, assuming that all currents are leaving except the current from current source, we have

$$(V_2-V_1)/3+ (V_2-V_1)/3+ (V_2-V_3)/2 = 5$$

$$-V_1[2/3]+V_2[1/3 +1/3 + 1/2]-V_3(1/2) =5$$

$$-0.66V_1+1.16V_2-0.5V_3= 5 \qquad (3.37)$$

At node 3 assuming all currents are leaving, we have

$(V_3-V_2)/2 + V_3/1 + V_3/6 =0$

$-0.5V_2 + 1.66V_3=0 \qquad (3.38)$

Applying Cramer's rule we get

$$V_1 = \frac{\begin{vmatrix} 1 & -0.66 & 0 \\ 5 & 1.16 & -0.5 \\ 0 & -0.5 & 1.66 \end{vmatrix}}{\begin{vmatrix} 0.96 & -0.66 & 0 \\ -0.66 & 1.16 & -0.5 \\ 0 & -0.5 & 1.66 \end{vmatrix}} = \frac{7.154}{0.887} = 8.06$$

Similarly,

$$V_2 = \frac{\begin{vmatrix} 0.96 & 1 & 0 \\ -0.66 & 5 & -0.5 \\ 0 & 0 & 1.66 \end{vmatrix}}{\begin{vmatrix} 0.96 & -0.66 & 0 \\ -0.66 & 1.16 & -0.5 \\ 0 & -0.5 & 1.66 \end{vmatrix}} = \frac{9.06}{0.887} = 10.2$$

$$V_3 = \frac{\begin{vmatrix} 0.96 & -0.66 & 1 \\ -0.66 & 1.16 & 5 \\ 0 & -0.5 & 0 \end{vmatrix}}{\begin{vmatrix} 0.96 & -0.66 & 0 \\ -0.66 & 1.16 & -0.5 \\ 0 & -0.5 & 1.66 \end{vmatrix}} = \frac{2.73}{0.887} = 3.07$$

3.6 NODAL EQUATIONS BY INSPECTION METHOD The nodal equations for a general planar network can also be written by inspection without going through the detailed steps. Consider a three node resistive network, including the reference node, as shown in fig 3.16

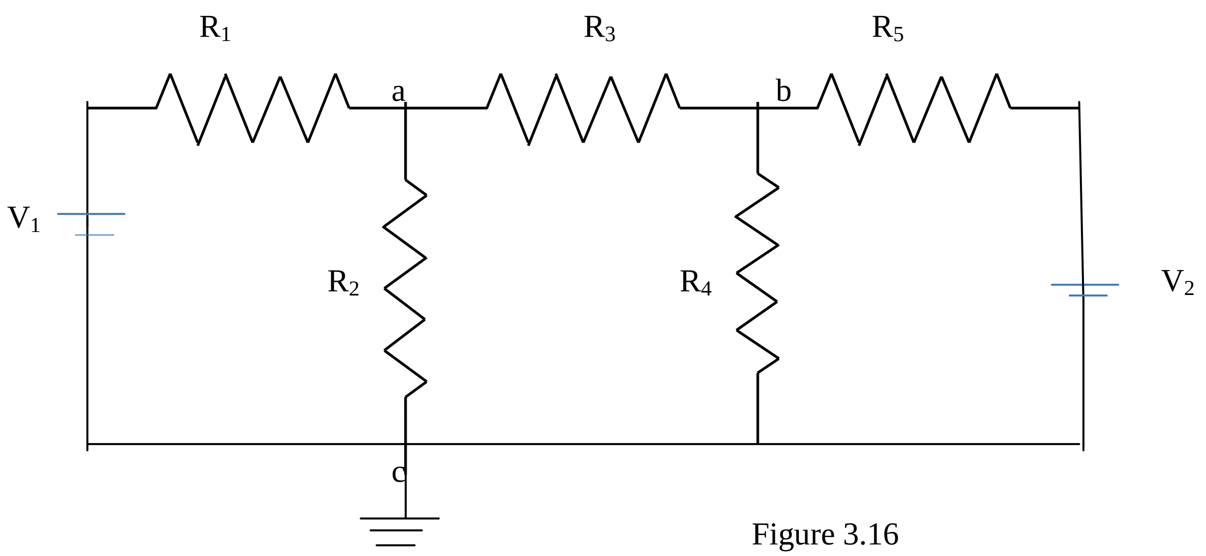


Figure 3.16

In fig. 3.16 the points a and b are the actual nodes and c is the reference node.

Now consider the nodes a and b separately as shown in fig 3.17(a) and (b)

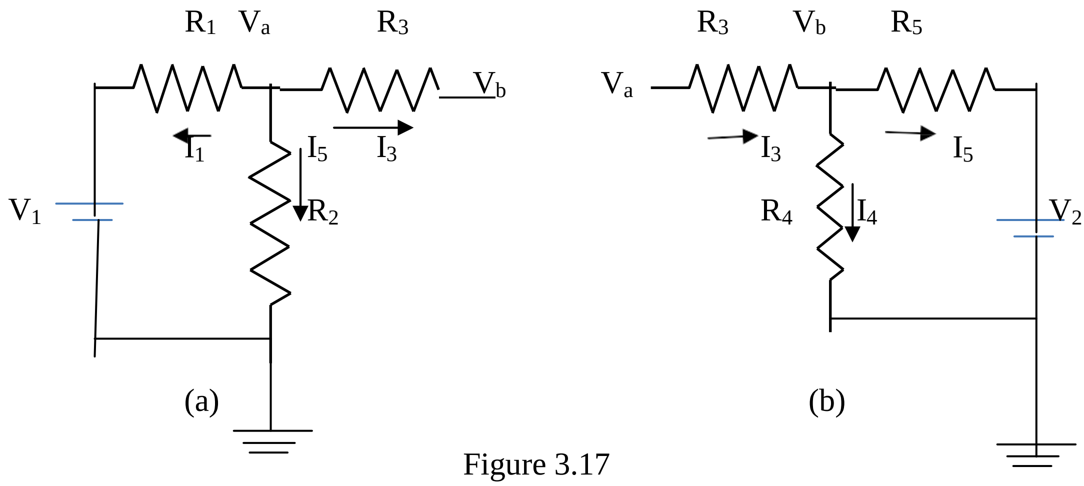


Figure 3.17

In fig 3.17 (a), according to Kirchhoff's current law we have

$I_1+I_2+I_3=0$

$(V_a-V_1)/R_1 + V_a/R_2 + (V_a-V_b)/R_3 = 0$ (3.39)

In fig 3.17 (b) , if we apply Kirchhoff's current law

$I_4 + I_5 = I_3$

$\therefore (V_b-V_a)/R_3 + V_b/R_4 + (V_b-V_2)/R_5 = 0$ (3.40)

Rearranging the above equations we get

$(1/R_1+1/R_2+1/R_3)V_a-(1/R_3)V_b=(1/R_1)V_1$ (3.41)

$(-1/R_3)V_a+ (1/R_3+1/R_4+1/R_5)V_b=V_2/R_5$ (3.42)

In general, the above equation can be written as

$G_{aa}V_a + G_{ab}V_b=I_1$ (3.43)

$G_{ba}V_a + G_{bb}V_b=I_2$ (3.44)

By comparing Eqs 3.41,3.42 and Eqs 3.43, 3.44 we have the self conductance at node a, $G_{aa}=(1/R_1 + 1/R_2 + 1/R_3)$ is the sum of the conductances connected to node a. Similarly, $G_{bb}= (1/R_3 + 1/R_4 +1/R_5)$ is the sum of the conductances connected to node b. $G_{ab}=(-1/R_3)$ is the sum of the mutual conductances connected to node *a* and node *b*. Here all the mutual conductances have negative signs. Similarly, $G_{ba}= (-1/R_3)$ is also a mutual conductance connected between nodes *b* and *a*. I_1 and I_2 are the sum of the source currents at node *a* and node *b*, respectively. The current which drives into the node has positive sign, while the current that drives away from the node has negative sign.

Example 3.8 for the circuit shown in the figure 3.18 write the node equations by the inspection method.

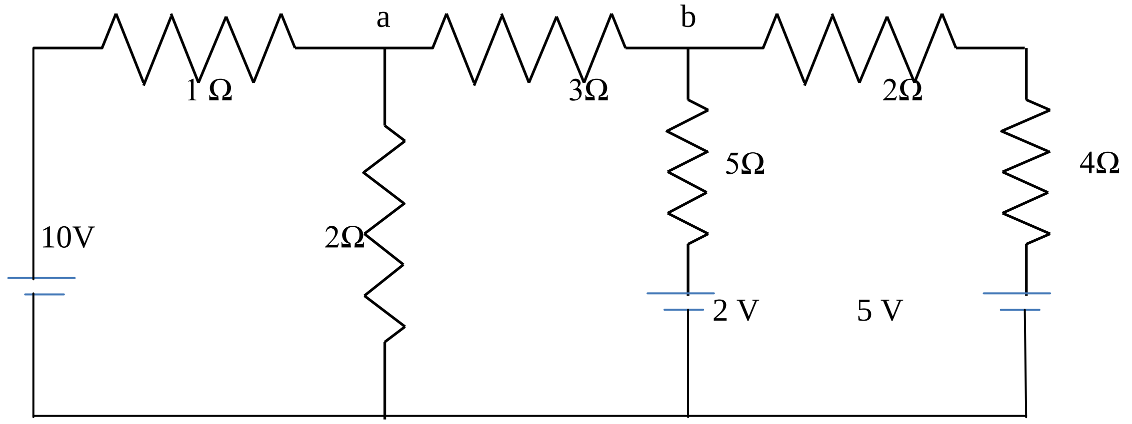


Fig 3.18

Solution:-

The general equations are

$G_{aa}V_a+G_{ab}V_b=I_1$ (3.45)

$G_{ba}V_a + G_{bb}V_b=I_2$ (3.46)

Consider equation 3.45

G_{aa}=(1+ 1/2 +1/3) mho. The self conductance at node *a* is the sum of the conductances connected to node *a*.

G_{bb} = (1/6 + 1/5 + 1/3) mho the self conductance at node *b* is the sum of conductances connected to node *b*.

G_{ab} =-(1/3) mho, the mutual conductances between nodes *a* and *b* is the sum of the conductances connected between node *a* and *b*.

Similarly G_{ba} = -(1/3), the sum of the mutual conductances between nodes *b* and *a*.

I_1=10/1 =10 A, the source current at node *a*,

I_2=(2/5 + 5/6) = 1.23A, the source current at node *b*.

Therefore, the nodal equations are

$1.83V_a - 0.33V_b = 10$ (3.47)

$-0.33V_a + 0.7V_b = 1.23$ (3.48)

3.7 SUPERNODE ANALYSIS

Suppose any of the branches in the network has a voltage source, then it is slightly difficult to apply nodal analysis. One way to overcome this difficulty is to apply the supernode technique. In this method, the two adjacent nodes that are connected by a voltage source are reduced to a single node and then the equations are formed by applying Kirchhoff's current law as usual. This is explained with the help of fig. 3.19

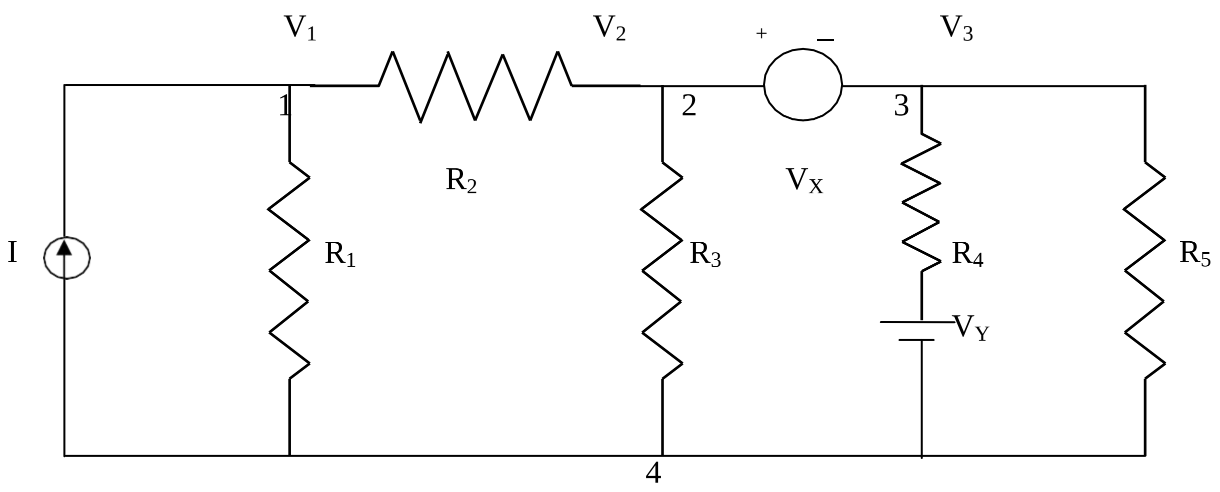


FIG 3.19

It is clear from the fig.3.19, that node 4 is the reference node. Applying Kirchhoff's current law at node 1, we get

$I = (V_1/R_1) + (V_1 - V_2)/R_2$

Due to the presence of voltage source V_χ in between nodes 2 and 3 , it is slightly difficult to find out the current. The supernode technique can be conveniently applied in this case.

Accordingly, we can write the combined equation for nodes 2 and 3 as under.

$(V_2-V_1)/R_2 + V_2/R_3 + (V_3-V_y)/R_4 + V_3/R_5 = 0$

The other equation is

$V_2-V_3 = V_x$

From the above three equations, we can find the three unknown voltages.

Example 3.9 Determine the current in the 5 Ω resistor for the circuit shown in fig. 3.20

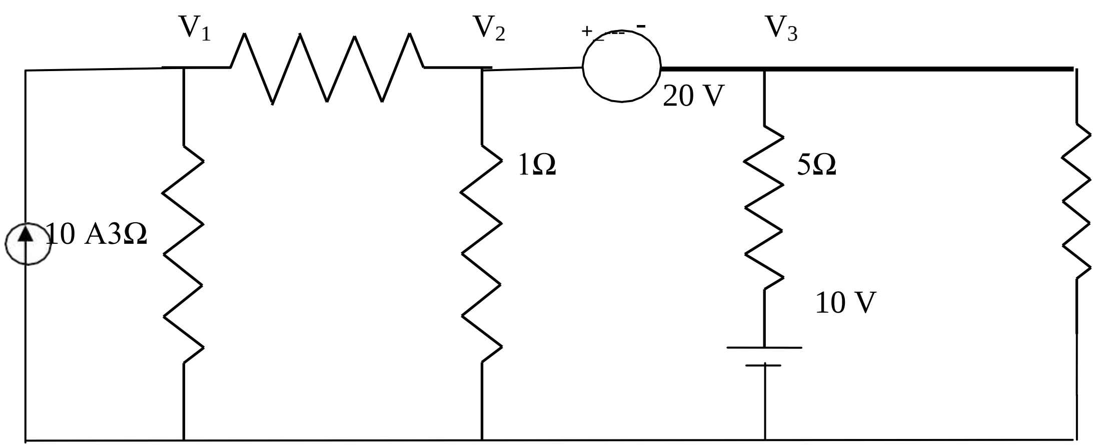


fig. 3.20

Solution. At node 1

$10 = V_1/3 + (V_1-V_2)/2$

Or $\quad V_1[1/3 + 1/2] - (V_2/2) - 10 = 0$

$0.83V_1 - 0.5V_2 - 10 = 0 \qquad (3.49)$

At node 2 and 3, the supernode equation is

$(V_2-V_1)/2 + V_2/1 + (V_3-10)/5 + V_3/2 = 0$

Or $\quad -V_1/2 + V_2[(1/2)+1] + V_3[1/5 + 1/2] = 2$

Or $\quad -0.5V_1 + 1.5V_2 + 0.7V_3 - 2 = 0 \qquad (2.50)$

The voltage between nodes 2 and 3 is given by

$V_2-V_3 = 20 \qquad (3.51)$

Chapter-04

NETWORK THEORM

INTRODUCTION

This chapter introduces a number of theorems that have application throughout the field of electricity and electronics. Not only can they be used to solve networks such as encountered in the previous chapter, but they also provide an opportunity to determine the impact of a particular source or element on the response of the entire system. In most cases, the network to be analyzed and the mathematics required to find the solution are simplified. All of the theorems appear again in the analysis of ac networks. In fact, the application of each theorem to ac networks is very similar in content to that found in this chapter.

The first theorem to be introduced is the superposition theorem, followed by Thévenin's theorem, Norton's theorem, and the maximum power transfer theorem. The chapter concludes with a brief introduction to Millman's theorem and the substitution and reciprocity theorems.

SUPERPOSITION THEOREM

The superposition theorem states that "The current through, or voltage across, any element of a network is equal to the algebraic sum of the currents or voltages produced independently by each source."

In other words, this theorem allows us to find a solution for a current or voltage using only one source at a time. Once we have the solution for each source, we can combine the results to obtain the total solution. The term algebraic appears in the above theorem statement because the currents resulting from the sources of the network can have different directions, just as the resulting voltages can have opposite polarities.

If we are to consider the effects of each source, the other sources obviously must be removed. Setting a voltage source to zero volts is like placing a short circuit across its terminals. Therefore, when removing a voltage source from a network schematic, replace it with a direct connection (short circuit) of zero ohms. Any internal resistance associated with the source must remain in the network.

Setting a current source to zero amperes is like replacing it with an open circuit. Therefore, when removing a current source from a network schematic, replace it by an open circuit of infinite ohms. Any internal resistance associated with the source must remain in the network.

The above statements are illustrated in Fig.

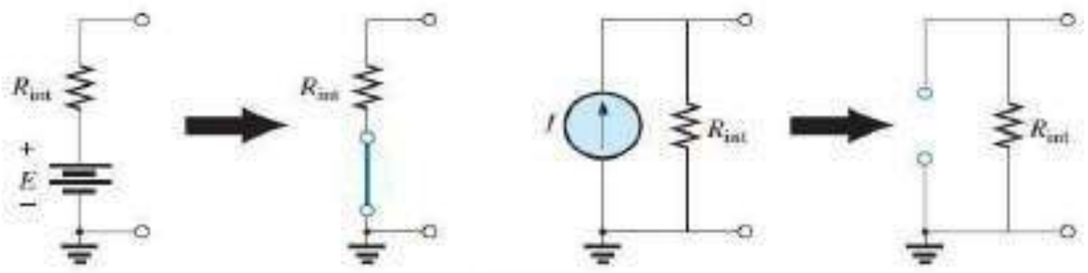


FIG. 9.1

Removing a voltage source and a current source to permit the application of the superposition theorem.

EXAMPLE 9.1

a. Using the superposition theorem, determine the current through resistor $R2$ for the network in Fig. 9.2.

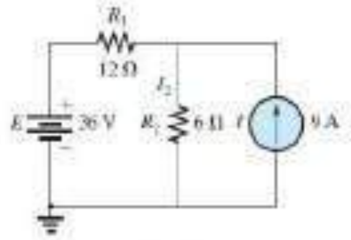


FIG. 9.2
Network to be analyzed in Example 9.1 using the superposition theorem.

Solutions:

In order to determine the effect of the 36 V voltage source, the current source must be replaced by an open-circuit equivalent as shown in Fig. 9.3. The result is a simple series circuit with a current equal to

$$I_2' = \frac{E}{R_T} = \frac{E}{R_1 + R_2} = \frac{36\text{ V}}{12\ \Omega + 6\ \Omega} = \frac{36\text{ V}}{18\ \Omega} = 2\text{ A}$$

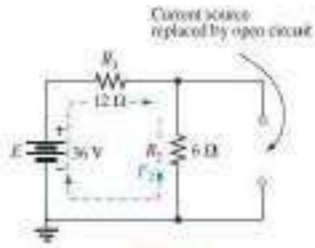


FIG. 9.3
Replacing the 9 A current source in Fig. 9.2 by an open circuit to determine the effect of the 36 V voltage source on current I_2.

Examining the effect of the 9 A current source requires replacing the 36 V voltage source by a short-circuit equivalent as shown in Fig. 9.4. The result is a parallel combination of resistors $R1$ and 2. Applying the current divider rule results in

$$I_2'' = \frac{R_1(I)}{R_1 + R_2} = \frac{(12\ \Omega)(9\text{ A})}{12\ \Omega + 6\ \Omega} = 6\text{ A}$$

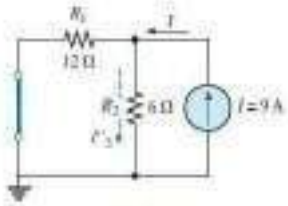


FIG. 9.4
Replacing the 36 V voltage source by a short-circuit equivalent to determine the effect of the 9 A current source on current I_2.

Since the contribution to current $I2$ has the same direction for each source, as shown in Fig. 9.5, the total solution for current $I2$ is the sum of the currents established by the two sources. That is,

$$I_2 = I_2' + I_2'' = 2\text{ A} + 6\text{ A} = \mathbf{8\text{ A}}$$

EXAMPLE 9.2 Using the superposition theorem, determine the current through the 12 Ω resistors in Fig. 9.8. Note that this is a two-source network of the type examined in the previous chapter when we applied branch-current analysis and mesh analysis.

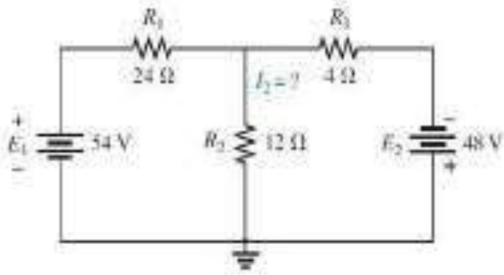


FIG. 9.8
Using the superposition theorem to determine the current through the 12 Ω resistor (Example 9.2).

Solution: Considering the effects of the 54 V source requires replacing the 48 V source by a short-circuit equivalent as shown in Fig. 9.9. The result is that the 12 Ω and 4 Ω resistors are in parallel. The total resistance seen by the source is therefore,

$$R_T = R_1 + R_2 \| R_3 = 24\ \Omega + 12\ \Omega \| 4\ \Omega = 24\ \Omega + 3\ \Omega = 27\ \Omega$$

and the source current is

$$I_s = \frac{E_1}{R_T} = \frac{54\text{ V}}{27\ \Omega} = 2\text{ A}$$

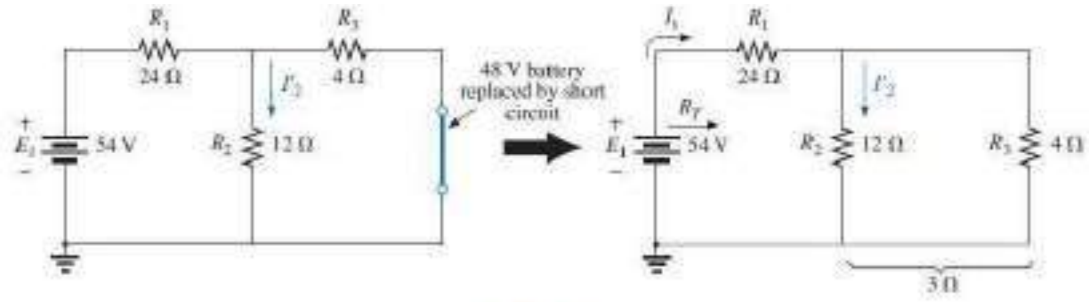


FIG. 9.9
Using the superposition theorem to determine the effect of the 54 V voltage source on current I_2 in Fig. 9.8.

Using the current divider rule results in the contribution to I_2 due to the 54 V source:

$$I'_2 = \frac{R_3 I_s}{R_3 + R_2} = \frac{(4\ \Omega)(2\ \text{A})}{4\ \Omega + 12\ \Omega} = 0.5\ \text{A}$$

If we now replace the 54 V source by a short-circuit equivalent, the network in Fig. 9.10 results. The result is a parallel connection for the 12 Ω and 24 Ω resistors.

Therefore, the total resistance seen by the 48 V source is

$$R_T = R_3 + R_2 \| R_1 = 4\ \Omega + 12\ \Omega \| 24\ \Omega = 4\ \Omega + 8\ \Omega = 12\ \Omega$$

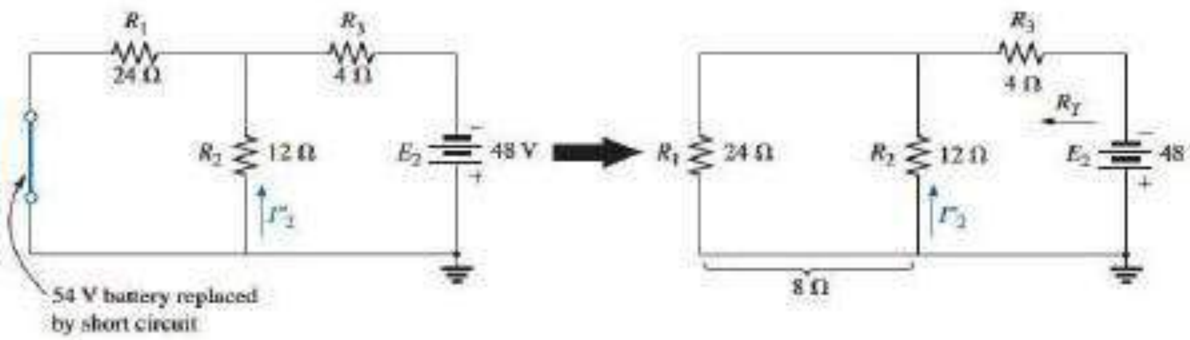


FIG. 9.10
Using the superposition theorem to determine the effect of the 48 V voltage source on current I_2 in Fig. 9.8.

and the source current is

$$I_s = \frac{E_2}{R_T} = \frac{48\ \text{V}}{12\ \Omega} = 4\ \text{A}$$

Applying the current divider rule results in

$$I''_2 = \frac{R_1(I_s)}{R_1 + R_2} = \frac{(24\ \Omega)(4\ \text{A})}{24\ \Omega + 12\ \Omega} = 2.67\ \text{A}$$

It is now important to realize that current I_2 due to each source has a different direction, as shown in Fig. 9.11. The net current therefore is the difference of the two and in the direction of the larger as follows:

$$I_2 = I''_2 - I'_2 = 2.67\ \text{A} - 0.5\ \text{A} = \mathbf{2.17\ A}$$

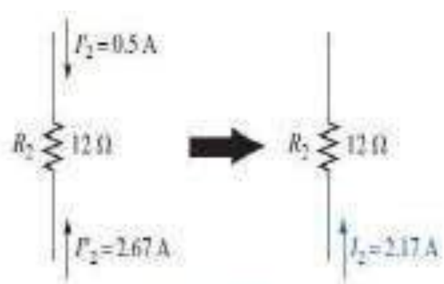


FIG. 9.11
Using the results of Figs. 9.9 and 9.10 to determine current I_2 for the network in Fig. 9.8.

EXAMPLE 9.3 Using the superposition theorem, determine current $I1$ for the network in

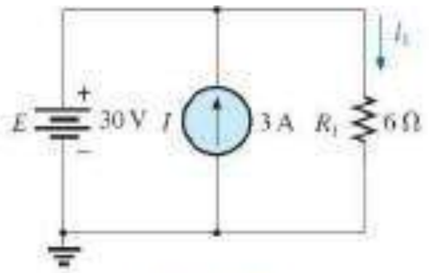


FIG. 9.12
Two-source network to be analyzed using the superposition theorem in Example 9.3.

Solution: Since two sources are present, there are two networks to be analyzed. First let us determine the effects of the voltage source by setting the current source to zero amperes as shown in Fig. 9.13. Note that the resulting current is defined as I_1' because it is the current through resistor $R1$ due to the voltage source only.

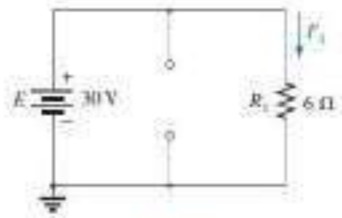


FIG. 9.13
Determining the effect of the 30 V supply on the current I_1 in Fig. 9.12.

Due to the open circuit, resistor $R1$ is in series (and, in fact, in parallel) with the voltage source E. The voltage across the resistor is the applied voltage, and current I_1' is determined by

$$I'_1 = \frac{V_1}{R_1} = \frac{E}{R_1} = \frac{30\ \text{V}}{6\ \Omega} = 5\ \text{A}$$

Now for the contribution due to the current source. Setting the voltage source to zero volts results in the network in Fig. 9.14, this presents us with an interesting situation. The current source has been replaced with a short-circuit equivalent that is directly across the current source and resistor $R1$. Since the source current takes the path of least resistance, it

chooses the zero ohm path of the inserted short-circuit equivalent, and the current through $R1$ is zero amperes. This is clearly demonstrated by an application of the current divider rule as follows:

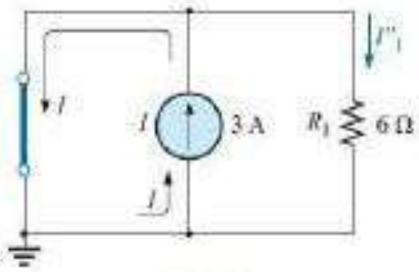


FIG. 9.14
Determining the effect of the 3 A current source on the current I_1 in Fig. 9.12.

$$I''_1 = \frac{R_{sc}I}{R_{sc} + R_1} = \frac{(0\ \Omega)I}{0\ \Omega + 6\ \Omega} = 0\ \text{A}$$

Since I'_1 and I''_1 have the same defined direction in Figs. 9.13 and 9.14, the total current is defined by

$$I_1 = I'_1 + I''_1 = 5\ \text{A} + 0\ \text{A} = 5\ \text{A}$$

Although this has been an excellent introduction to the application of the superposition theorem, it should be immediately clear in Fig. 9.12 that the voltage source is in parallel with the current source and load resistor R_1, so the voltage across each must be 30 V. The result is that I_1 must be determined solely by

$$I_1 = \frac{V_1}{R_1} = \frac{E}{R_1} = \frac{30\ \text{V}}{6\ \Omega} = 5\ \text{A}$$

EXAMPLE 9.4 Using the principle of superposition, find the current I_2 through the 12 kΩ resistor in Fig. 9.15.

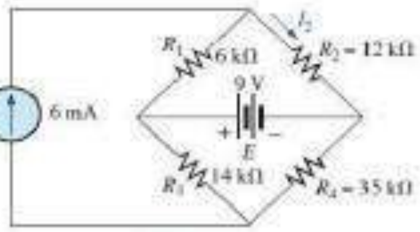


FIG. 9.15
Example 9.4.

Solution: Consider the effect of the 6 mA current source (Fig. 9.16).

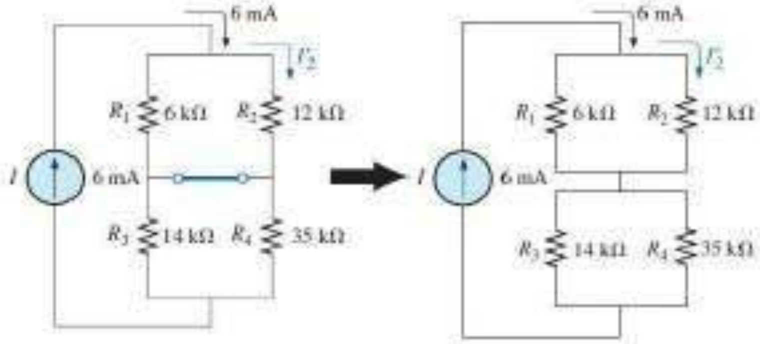


FIG. 9.16
The effect of the current source I on the current I_2.

The current divider rule gives

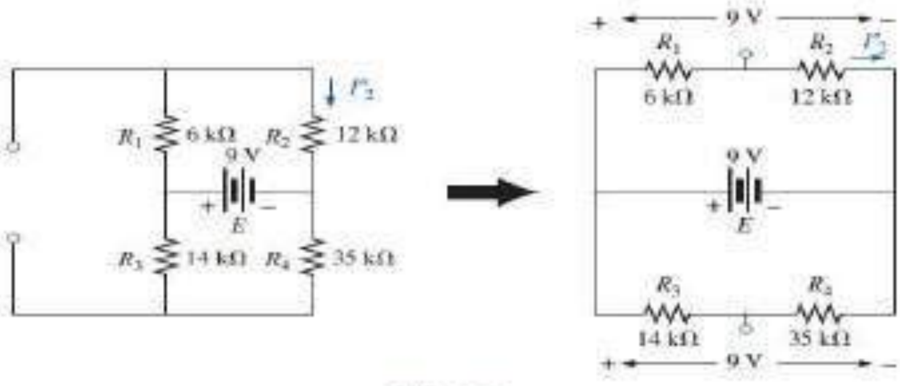
$$I'_2 = \frac{R_1 I}{R_1 + R_2} = \frac{(6\ \text{k}\Omega)(6\ \text{mA})}{6\ \text{k}\Omega + 12\ \text{k}\Omega} = 2\ \text{mA}$$

Considering the effect of the 9 V voltage source (Fig. 9.17) gives

$$I''_2 = \frac{E}{R_1 + R_2} = \frac{9\ \text{V}}{6\ \text{k}\Omega + 12\ \text{k}\Omega} = 0.5\ \text{mA}$$

Since I'_2 and I''_2 have the same direction through R_2, the desired current is the sum of the two:

$$\begin{aligned} I_2 &= I'_2 + I''_2 \\ &= 2\ \text{mA} + 0.5\ \text{mA} \\ &= \mathbf{2.5\ mA} \end{aligned}$$



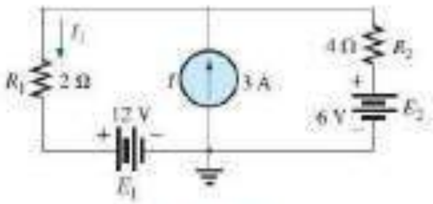


FIG. 9.18
Example 9.5.

EXAMPLE 9.5 Find the current through the 2 Ω resistor of the network in Fig. 9.18. The presence of three sources results in three different networks to be analyzed.

Solution: Consider the effect of the 12 V source (Fig. 9.19):

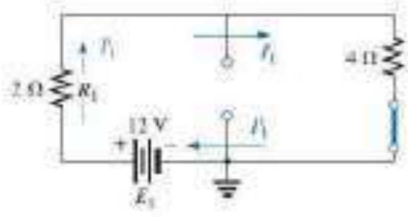


FIG. 9.19
The effect of E_1 on the current I.

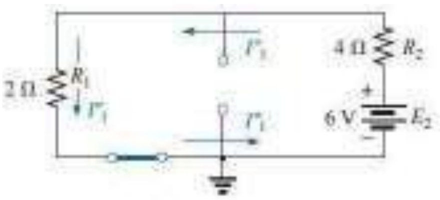


FIG. 9.20
The effect of E_2 on the current I_1.

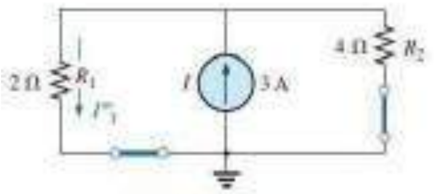


FIG. 9.21
The effect of I on the current I_1.

$$I'_1 = \frac{E_1}{R_1 + R_2} = \frac{12\text{ V}}{2\ \Omega + 4\ \Omega} = \frac{12\text{ V}}{6\ \Omega} = 2\text{ A}$$

Consider the effect of the 6 V source (Fig. 9.20):

$$I''_1 = \frac{E_2}{R_1 + R_2} = \frac{6\text{ V}}{2\ \Omega + 4\ \Omega} = \frac{6\text{ V}}{6\ \Omega} = 1\text{ A}$$

Consider the effect of the 3 A source (Fig. 9.21): Applying the current divider rule gives

$$I'''_1 = \frac{R_2 I}{R_1 + R_2} = \frac{(4\ \Omega)(3\text{ A})}{2\ \Omega + 4\ \Omega} = \frac{12\text{ A}}{6} = 2\text{ A}$$

The total current through the 2 Ω resistor appears in Fig. 9.22, and

$$\begin{aligned} I_1 &= \overbrace{I''_1 + I'''_1}^{\text{Same direction as } I_1 \text{ in Fig. 9.18}} \overset{\text{Opposite direction to } I_1 \text{ in Fig. 9.18}}{-} I'_1 \\ &= 1\text{ A} + 2\text{ A} - 2\text{ A} = 1\text{ A} \end{aligned}$$

9.3 THÉVENIN'S THEOREM

The next theorem to be introduced, **Thévenin's theorem,** is probably one of the most interesting in that it permits the reduction of complex networks to a simpler form for analysis and design.

In general, the theorem can be used to do the following:

- ***Analyze networks with sources that are not in series or parallel.***
- ***Reduce the number of components required to establish the same characteristics at the output terminals.***
- ***Investigate the effect of changing a particular component on the behaviour of a network without having to analyze the entire network after each change.***

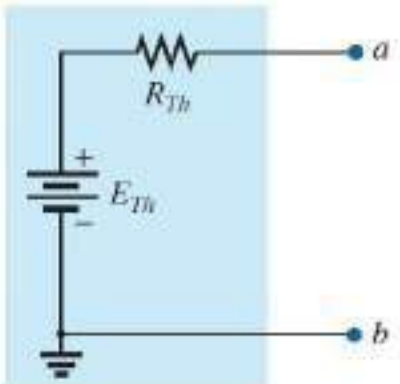


FIG. 9.23
Thévenin equivalent circuit.

All three areas of application are demonstrated in the examples to follow.

Thévenin's theorem states the following:

Any two-terminal dc network can be replaced by an equivalent circuit consisting solely of a voltage source and a series resistor as shown in Fig. 9.23.

The theorem was developed by Commandant Leon-Charles Thévenin in 1883 as described in Fig. 9.24.

To demonstrate the power of the theorem, consider the fairly complex network of Fig. 9.25(a) with its two sources and series-parallel connections.

The theorem states that the entire network inside the blue shaded area can be replaced by one voltage source and one resistor as shown in Fig. 9.25(b). If the replacement is done properly, the voltage across, and the current through, the resistor RL will be the same for each network. The value of RL can be changed to any

FIG. 9.24
Leon-Charles Thévenin.
Courtesy of the Bibliothèque École Polytechnique, Paris, France.

value, and the voltage, current, or power to the load resistor is the same for each configuration.

Now, this is a very powerful statement—one that is verified in the examples to follow.

The question then is, How can you determine the proper value of Thévenin voltage and resistance? In general, finding the Thévenin *resistance* value is quite straightforward. Finding the Thévenin *voltage* can be more of a challenge and, in fact, may require using the superposition theorem.

Fortunately, there is a series of steps that will lead to the proper value of each parameter. Although a few of the steps may seem trivial at first, they can become quite important when the network becomes complex.

Thévenin's Theorem Procedure

Preliminary:

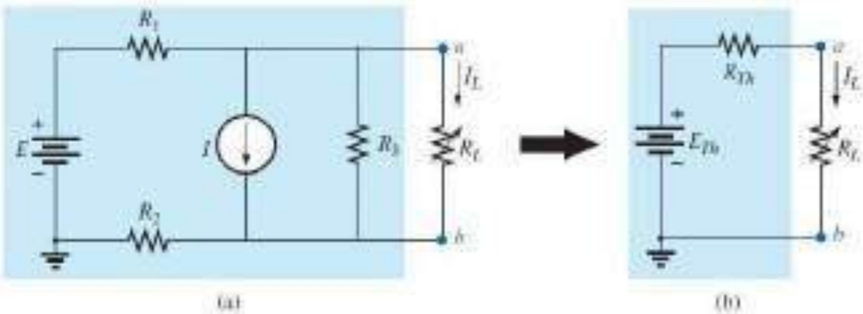
1. Remove that portion of the network where the Thévenin equivalent circuit is found. In Fig. 9.25(a), this requires that the load resistor RL be temporarily removed from the network.

2. Mark the terminals of the remaining two-terminal network. (The importance of this step will become obvious as we progress through some complex networks.) RTh:

3. Calculate RTh by first setting all sources to zero (voltage sources are replaced by short circuits and current sources by open circuits) and then finding the resultant resistance between the two marked terminals. (If the internal resistance of the voltage and/or current sources is included in the original network, it must remain when the sources are set to zero.) ETh:

4. Calculate ETh by first returning all sources to their original position and finding the open-circuit voltage between the marked terminals. (This step is invariably the one that causes most confusion and errors. In all cases, keep in mind that it is the open-circuit potential between the two terminals marked in step 2.) Conclusion:

5. Draw the Thévenin equivalent circuit with the portion of the circuit previously removed replaced between the terminals of the equivalent circuit. This step is indicated by the placement of the resistor RL between the terminals of the Thévenin equivalent circuit as shown in Fig. 9.25(b).



(a) (b)

FIG. 9.25
Substituting the Thévenin equivalent circuit for a complex network.

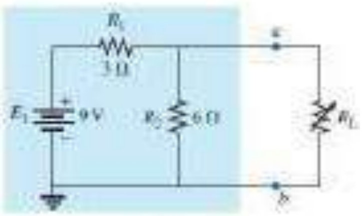


FIG. 9.26
Example 9.6.

EXAMPLE 9.6 Find the Thévenin equivalent circuit for the network in the shaded area of the network in Fig. 9.26. Then find the current through R_L for values of 2 Ω, 10 Ω, and 100 Ω.

Solution:

Steps 1 and 2: These produce the network in Fig. 9.27. Note that the load resistor R_L has been removed and the two "holding" terminals have been defined as *a* and *b*.

Step 3: Replacing the voltage source E_1 with a short-circuit equivalent yields the network in Fig. 9.28(a), where

$$R_{Th} = R_1 \| R_2 = \frac{(3\ \Omega)(6\ \Omega)}{3\ \Omega + 6\ \Omega} = \mathbf{2\ \Omega}$$

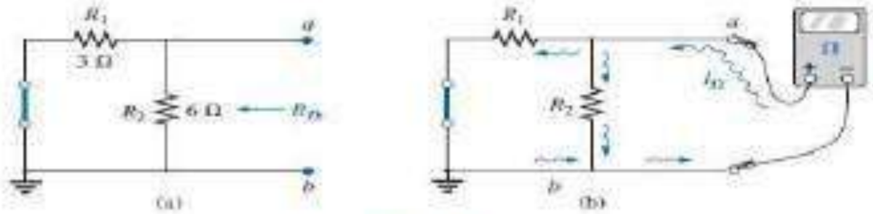


FIG. 9.28
Determining R_{Th} for the network in Fig. 9.27.

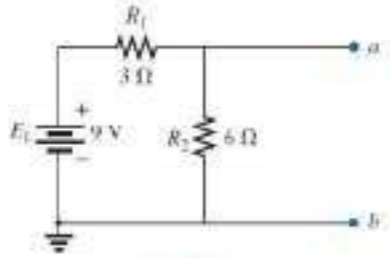


FIG. 9.27
Identifying the terminals of particular importance when applying Thévenin's theorem.

The importance of the two marked terminals now begins to surface. They are the two terminals across which the Thévenin resistance is measured. It is no longer the total resistance as seen by the source, as determined in the majority of problems of Chapter 7. If some difficulty develops when determining R_{Th} with regard to whether the resistive elements are in series or parallel, consider recalling that the ohmmeter sends out a trickle current into a resistive combination and senses the level of the resulting voltage to establish the measured resistance level. In Fig. 9.28(b), the trickle current of the ohmmeter approaches the network through terminal *a*, and when it reaches the junction of *R*1 and *R*2, it splits as shown. The fact that the trickle current splits and then recombines at the lower node reveals that the resistors are in parallel as far as the ohmmeter reading is concerned. In essence, the path of the sensing current of the ohmmeter has revealed how the resistors are connected to the two terminals of interest and how the Thévenin resistance should be determined. Remember this as you work through the various examples in this section.

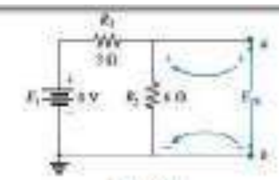


FIG. 9.29
Determining E_{Th} for the network in Fig. 9.27.

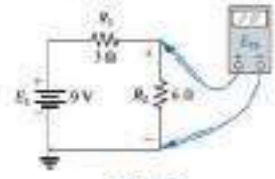


FIG. 9.30
Measuring E_{Th} for the network in Fig. 9.27.

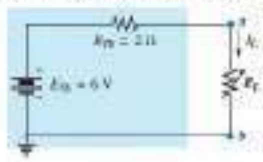


FIG. 9.31
Substituting the Thévenin equivalent circuit for the network external to R_L in Fig. 9.26.

Step 4: Replace the voltage source (Fig. 9.29). For this case, the opencircuit voltage *ETh* is the same as the voltage drop across the 6 Ω resistor.

Applying the voltage divider rule gives

$$E_{Th} = \frac{R_2 E_1}{R_2 + R_1} = \frac{(6\ \Omega)(9\ \text{V})}{6\ \Omega + 3\ \Omega} = \frac{54\ \text{V}}{9} = \mathbf{6\ V}$$

It is particularly important to recognize that E_{Th} is the open-circuit potential between points *a* and *b*. Remember that an open circuit can have any voltage across it, but the current must be zero. In fact, the current through any element in series with the open circuit must be zero also. The use of a voltmeter to measure E_{Th} appears in Fig. 9.30. Note that it is placed directly across the resistor R_2 since E_{Th} and V_{R_2} are in parallel.

Step 5: (Fig. 9.31):

$$I_L = \frac{E_{Th}}{R_{Th} + R_L}$$

$$R_L = 2\ \Omega: \qquad I_L = \frac{6\ \text{V}}{2\ \Omega + 2\ \Omega} = \mathbf{1.5\ A}$$

$$R_L = 10\ \Omega: \qquad I_L = \frac{6\ \text{V}}{2\ \Omega + 10\ \Omega} = \mathbf{0.5\ A}$$

$$R_L = 100\ \Omega: \qquad I_L = \frac{6\ \text{V}}{2\ \Omega + 100\ \Omega} = \mathbf{0.06\ A}$$

If Thévenin's theorem were unavailable, each change in R_L would require that the entire network in Fig. 9.26 be reexamined to find the new value of R_L.

EXAMPLE 9.7 Find the Thévenin equivalent circuit for the network in the shaded area of the network in Fig. 9.32.

Solution:

Steps 1 and 2: See Fig. 9.33.

Step 3: See Fig. 9.34. The current source has been replaced with an open-circuit equivalent and the resistance determined between terminals *a* and *b*.

In this case, an ohmmeter connected between terminals *a* and *b* sends out a sensing current that flows directly through *R*1 and *R*2 (at the same level). The result is that *R*1 and *R*2 are in series and the Thévenin resistance is the sum of the two,

$$R_{Th} = R_1 + R_2 = 4\ \Omega + 2\ \Omega = \mathbf{6\ \Omega}$$

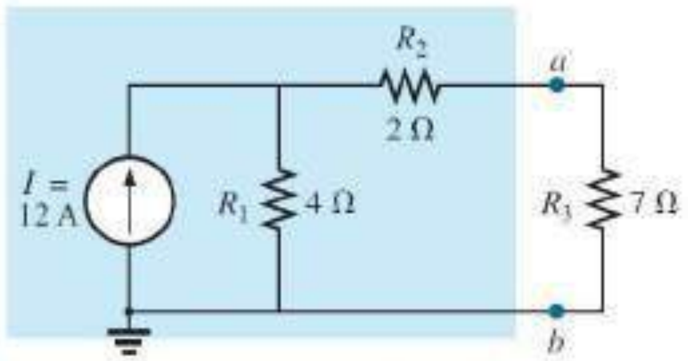


FIG. 9.32
Example 9.7.

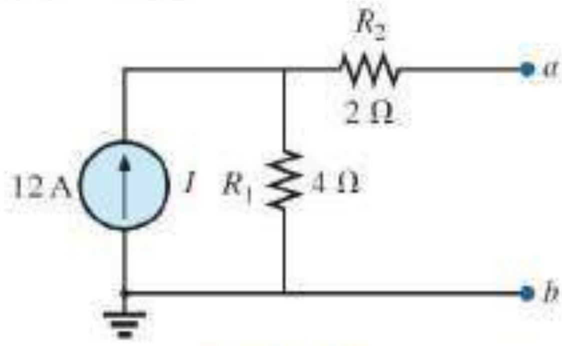


FIG. 9.33
Establishing the terminals of particular interest for the network in Fig. 9.32.

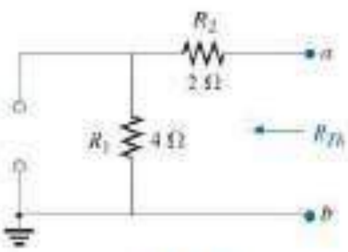


FIG. 9.34
Determining R_{Th} for the network in Fig. 9.33.

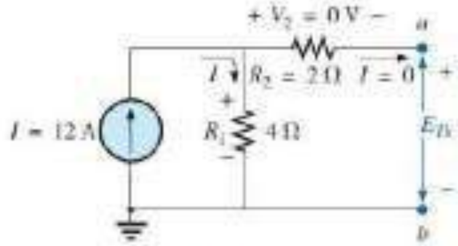


FIG. 9.35
Determining E_{Th} for the network in Fig. 9.33.

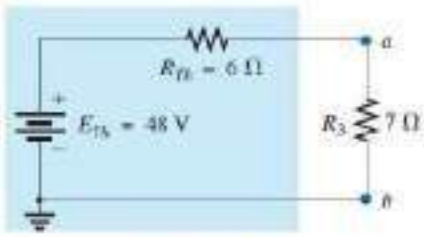


FIG. 9.36
Substituting the Thévenin equivalent circuit in the network external to the resistor R_3 in Fig. 9.32.

Step 4: See Fig. 9.35. In this case, since an open circuit exists between the two marked terminals, the current is zero between these terminals and through the 2Ω resistor. The voltage drop across *R*2 is, therefore,

$$V2 = I2R2 = (0)\ R2 = 0\ \text{V}$$

and $$ETh = V1 = I1R1 = IR1 = (12\ \text{A})(4\ \_) = \mathbf{48\ V}$$

Step 5: See Fig. 9.36.

EXAMPLE 9.8 Find the Thévenin equivalent circuit for the network in the shaded area of the network in Fig. 9.37. Note in this example that there is no need for the section of the network to be preserved to be at the "end" of the configuration.

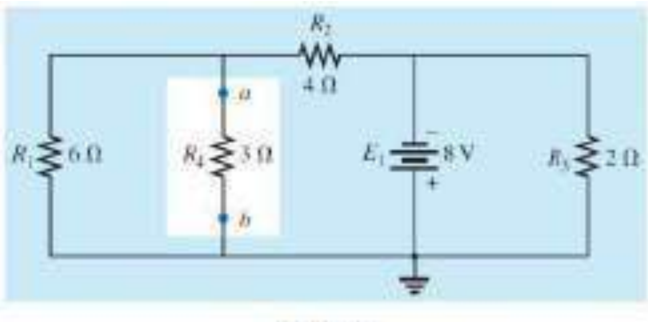


FIG. 9.37

Solution:

Steps 1 and 2: See Fig. 9.38

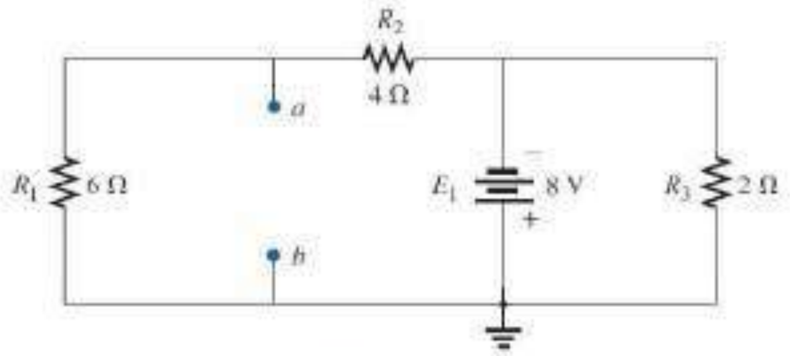


FIG. 9.38

Identifying the terminals of particular interest for the network in Fig. 9.37.

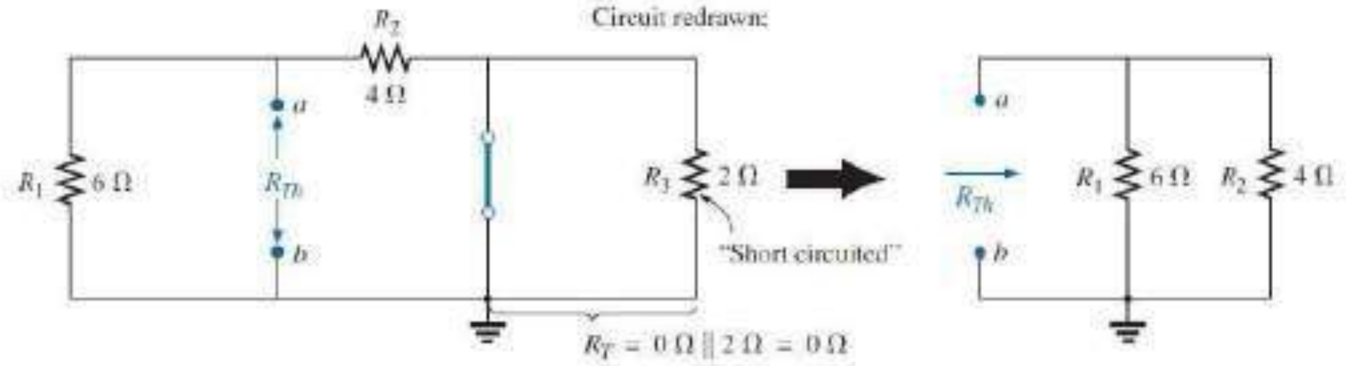


FIG. 9.39

Determining R_{Th} for the network in Fig. 9.38.

Step 3: See Fig. 9.39. Steps 1 and 2 are relatively easy to apply, but now we must be careful to "hold" onto the terminals *a* and *b* as the Thévenin resistance and voltage are determined. In Fig. 9.39, all the remaining elements turn out to be in parallel, and the network can be redrawn as shown. We have

$$R_{Th} = R_1 \| R_2 = \frac{(6\ \Omega)(4\ \Omega)}{6\ \Omega + 4\ \Omega} = \frac{24\ \Omega}{10} = \mathbf{2.4\ \Omega}$$

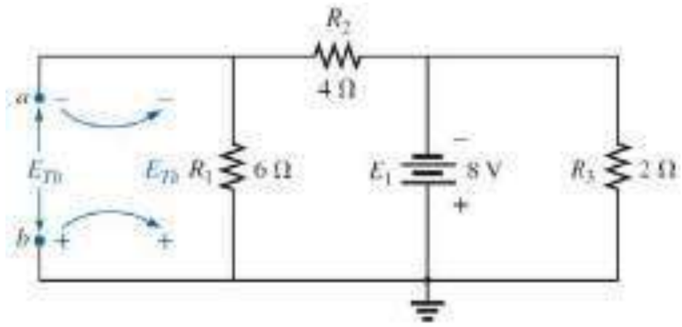


FIG. 9.40

Determining E_{Th} for the network in Fig. 9.38.

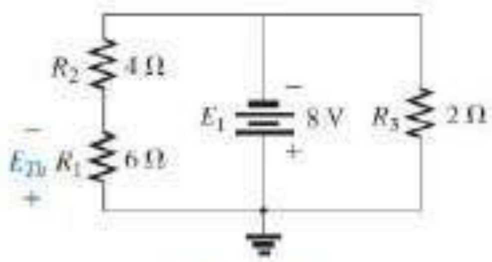


FIG. 9.41

Network of Fig. 9.40 redrawn.

Step 4: See Fig. 9.40. In this case, the network can be redrawn as shown in Fig. 9.41. Since the voltage is the same across parallel elements, the voltage across the series resistors *R*1 and *R*2 is *E*1, or 8 V. Applying the voltage divider rule gives

$$E_{Th} = \frac{R_1 E_1}{R_1 + R_2} = \frac{(6\ \Omega)(8\ \text{V})}{6\ \Omega + 4\ \Omega} = \frac{48\ \text{V}}{10} = \mathbf{4.8\ V}$$

Step 5: See Fig. 9.42.

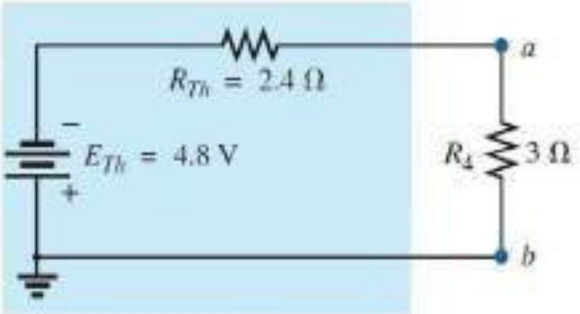


FIG. 9.42
Substituting the Thévenin equivalent circuit for the network external to the resistor R_4 in Fig. 9.37.

EXAMPLE 9.9 Find the Thévenin equivalent circuit for the network in the shaded area of the bridge network in Fig. 9.43.

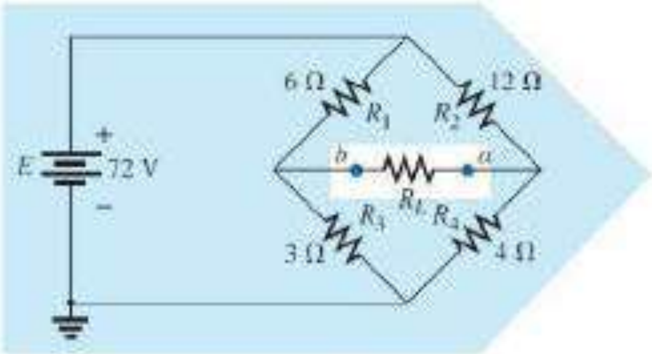


FIG. 9.43
Example 9.9.

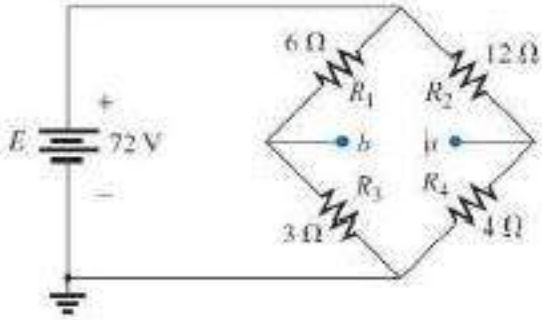


FIG. 9.44
Identifying the terminals of particular interest for the network in Fig. 9.43.

Solution:

Steps 1 and 2: See Fig. 9.44.

Step 3: See Fig. 9.45. In this case, the short-circuit replacement of the voltage source *E* provides a direct connection between *c* and *c*_ in Fig. 9.45(a), permitting a “folding” of the network around the horizontal line of *a*-*b* to produce the configuration in Fig. 9.45(b).

$$R_{Th} = R_{a-b} = R_1 \| R_3 + R_2 \| R_4$$
$$= 6\ \Omega \| 3\ \Omega + 4\ \Omega \| 12\ \Omega$$
$$= 2\ \Omega + 3\ \Omega = \mathbf{5\ \Omega}$$

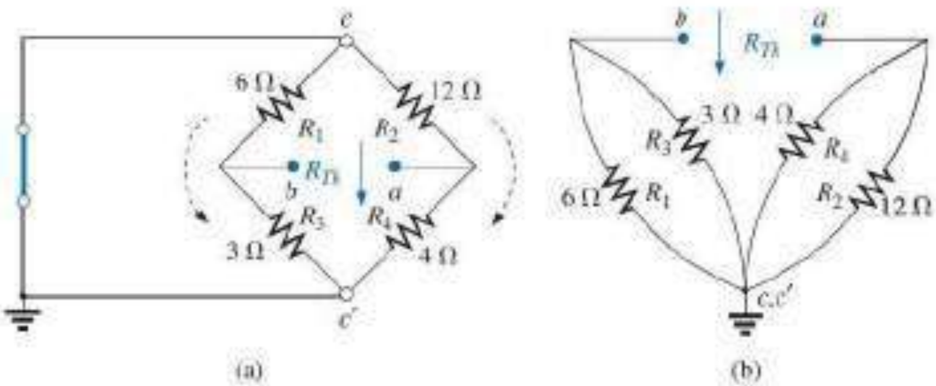


FIG. 9.45
Solving for R_{Th} for the network in Fig. 9.44.

Step 4: The circuit is redrawn in Fig. 9.46. The absence of a direct connection between *a* and *b* results in a network with three parallel branches. The voltages V_1 and V_2 can therefore be determined using the voltage divider rule:

$$V_1 = \frac{R_1E}{R_1 + R_3} = \frac{(6\ \Omega)(72\ \text{V})}{6\ \Omega + 3\ \Omega} = \frac{432\ \text{V}}{9} = 48\ \text{V}$$

$$V_2 = \frac{R_2E}{R_2 + R_4} = \frac{(12\ \Omega)(72\ \text{V})}{12\ \Omega + 4\ \Omega} = \frac{864\ \text{V}}{16} = 54\ \text{V}$$

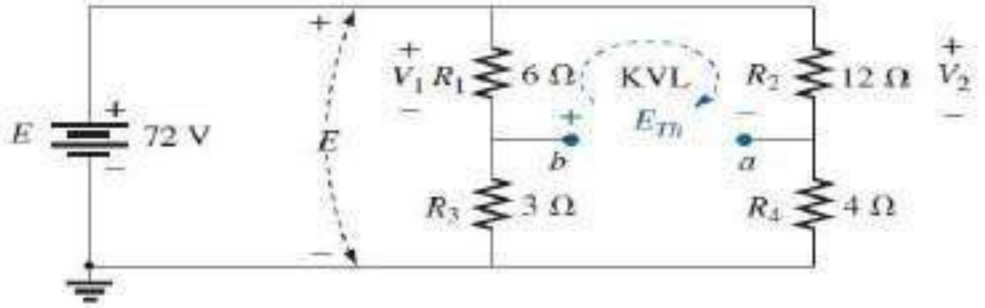


FIG. 9.46
Determining E_{Th} for the network in Fig. 9.44.

Assuming the polarity shown for E_{Th} and applying Kirchhoff's voltage law to the top loop in the clockwise direction results in

$$\Sigma_C V = +E_{Th} + V_1 - V_2 = 0$$

and
$$E_{Th} = V_2 - V_1 = 54\ \text{V} - 48\ \text{V} = \mathbf{6\ V}$$

Step 5: See Fig. 9.47.

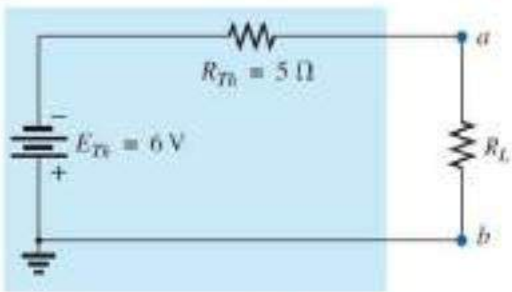


FIG. 9.47

EXAMPLE 9.11 For the network of Fig. 9.54,
a. Find the Thévenin equivalent circuit for the portion of the network in the shaded area.
b. Reconstruct the network of Fig. 9.54 with the Thévenin equivalent network in place.
c. Using the resulting network of part (b) find the voltage *Va*.

Solutions:

a. *Steps 1 and 2:* See Fig. 9.55.
Step 3: See Fig. 9.56.

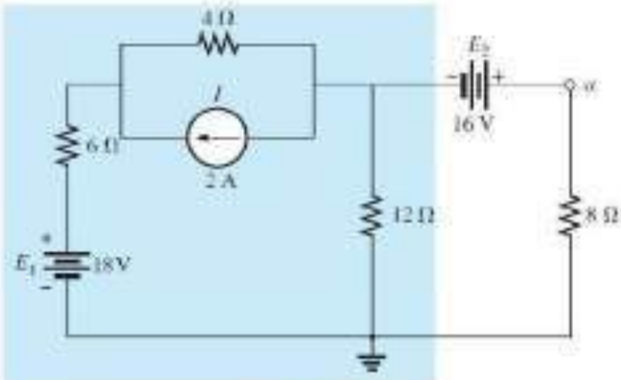


FIG. 9.54
Example 9.11.

$$R_{Th} = 12\ \Omega \| (4\ \Omega + 6\ \Omega) = 12\ \Omega \| 10\ \Omega = 5$$

Step 4: Applying the superposition theorem, we will first find the effect of the voltage source on the Thévenin voltage using the network of Fig. 9.57. Applying the voltage divider rule:

$$E'_{Th} = \frac{12\ \Omega\ (18\ \text{V})}{6\ \Omega + 4\ \Omega + 12\ \Omega} = \frac{216}{22}\text{V} = 9.82\ \text{V}$$

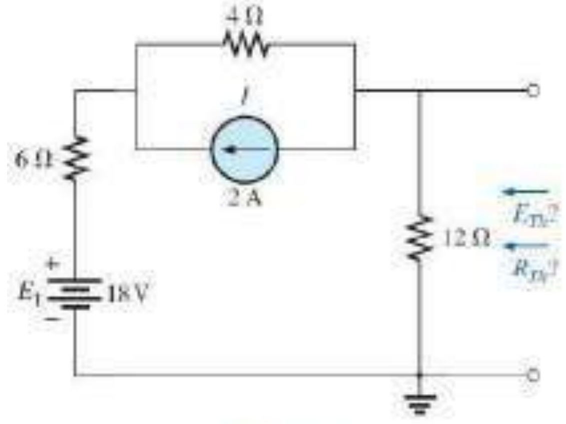


FIG. 9.55
Establishing the terminals of interest for the network of Fig. 9.54.

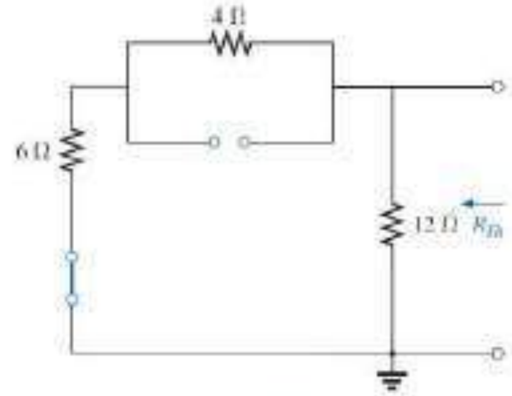
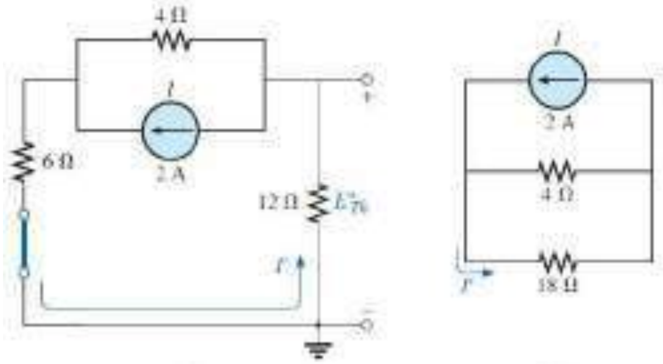


FIG. 9.56
Determining R_{Th}.



(a) (b)

FIG. 9.58
Determining the contribution of I to E_{Th}.

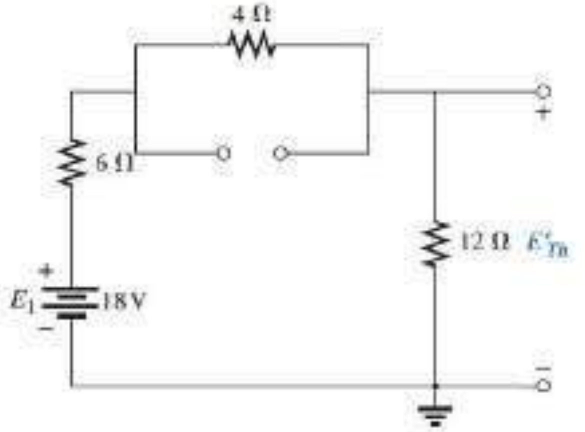


FIG. 9.57
Determining the contribution of E_1 to E_{Th}.

The contribution due to the current source is determined using the network of Fig. 9.58(a) redrawn as shown in Fig. 9.58(b). Applying the current divider rule:

$$I'' = \frac{4\ \Omega\,(2\ \text{A})}{4\ \Omega + 18\ \Omega} = \frac{8}{22}\ \text{A} = 0.364\ \text{A}$$

and $\quad E''_{Th} = -I''(12\ \Omega) = -(0.364\ \text{A})(12\ \Omega) = -4.37\ \text{V}$

so that $\quad E_{Th} = E'_{Th} + E''_{Th} = 9.82\ \text{V} - 4.37\ \text{V} = \mathbf{5.45\ V}$

b. The reconstructed network is shown in Fig. 9.59.
c. Using the voltage divider rule:

$$V_a = \frac{8\ \Omega\,(5.45\ \text{V} + 16\ \text{V})}{5.45\ \Omega + 8\ \Omega} = \frac{8\,(21.45)}{13.45}\ \text{V} = \frac{171.6}{13.45}\ \text{V} = \mathbf{12.76\ V}$$

Instead of using the superposition theorem, the current source could first have been converted to a voltage source and the series elements combined to determine the Thévenin voltage. In any event both approaches would have yielded the same results.

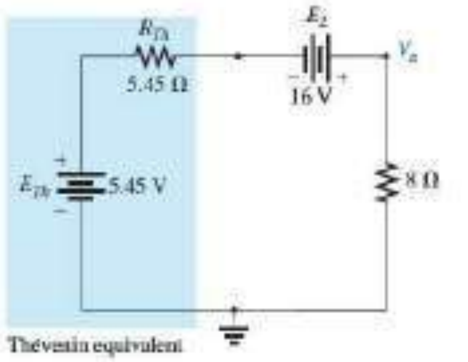


FIG. 9.59
Applying the Thévenin equivalent network to the network of Fig. 9.54.

9.4 NORTON'S THEOREM

Any two-terminal linear bilateral dc network can be replaced by an
equivalent circuit consisting of a current source and a parallel
resistor, as shown in Fig. 9.65.

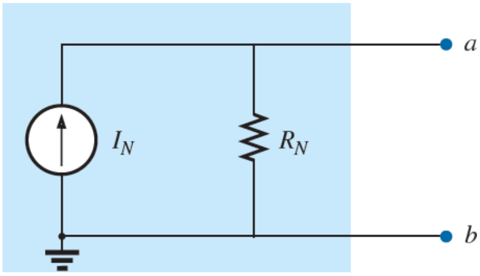


FIG. 9.65
Norton equivalent circuit.

FIG. 9.64
Edward L. Norton.
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The discussion of Thévenin's theorem with respect to the equivalent circuit can also be applied to the Norton equivalent circuit. The steps leading to the proper values of *IN* and *RN* are now listed.

Norton's Theorem Procedure

Preliminary:

1. Remove that portion of the network across which the Norton equivalent circuit is found.

2. Mark the terminals of the remaining two-terminal network.

RN:

3. Calculate RN by first setting all sources to zero (voltage sources are replaced with short circuits and current sources with open circuits) and then finding the resultant resistance between the two marked terminals. (If the internal resistance of the voltage and/or current sources is included in the original network, it must remain when the sources are set to zero.) Since RN _ RTh, the procedure and value obtained using the approach described for Thévenin's theorem will determine the proper value of RN.

IN:

4. Calculate IN by first returning all sources to their original position and then finding the short-circuit current between the marked terminals. It is the same current that would be measured by an ammeter placed between the marked terminals.

Conclusion:

5. Draw the Norton equivalent circuit with the portion of the circuit previously removed replaced between the terminals of the equivalent circuit.

The Norton and Thévenin equivalent circuits can also be found from each other by using the source transformation discussed earlier in this chapter and reproduced in Fig. 9.66.

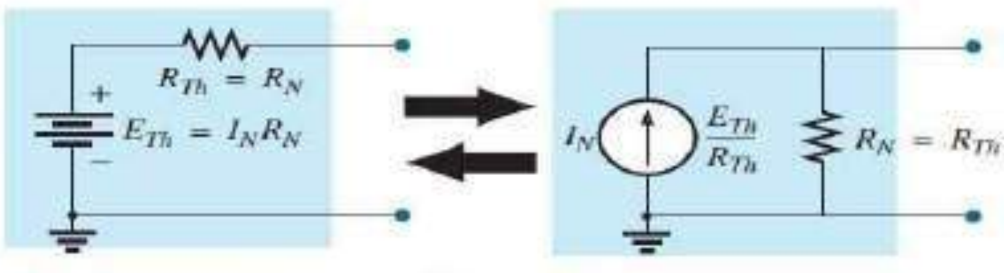


FIG. 9.66
Converting between Thévenin and Norton equivalent circuits.

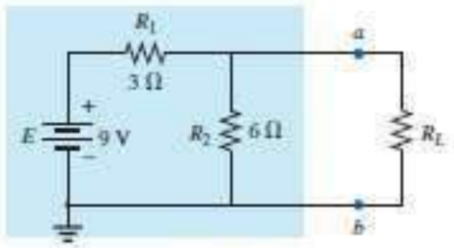


FIG. 9.67
Example 9.12.

EXAMPLE 9.12 Find the Norton equivalent circuit for the network in the shaded area in Fig. 9.67.

Solution:

Steps 1 and 2: See Fig. 9.68.

Step 3: See Fig. 9.69, and

$$R_N = R_1 \| R_2 = 3\ \Omega \| 6\ \Omega = \frac{(3\ \Omega)(6\ \Omega)}{3\ \Omega + 6\ \Omega} = \frac{18\ \Omega}{9} = \mathbf{2\ \Omega}$$

Step 4: See Fig. 9.70, which clearly indicates that the short-circuit connection between terminals *a* and *b* is in parallel with *R*2 and eliminates its effect. *IN* is therefore the same as through *R*1, and the full battery voltage appears across *R*1 since

$$V_2 = I_2R_2 = (0)6\ \Omega = 0\ \text{V}$$

Therefore,

$$I_N = \frac{E}{R_1} = \frac{9\ \text{V}}{3\ \Omega} = \mathbf{3\ A}$$

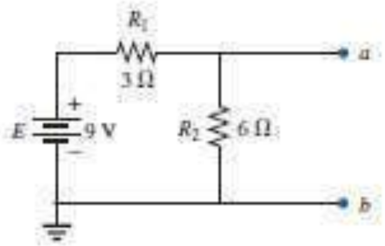


FIG. 9.68
Identifying the terminals of particular interest for the network in Fig. 9.67.

Step 5: See Fig. 9.71. This circuit is the same as the first one considered in the development of Thévenin's theorem. A simple conversion indicates that the Thévenin circuits are, in fact, the same (Fig. 9.72).

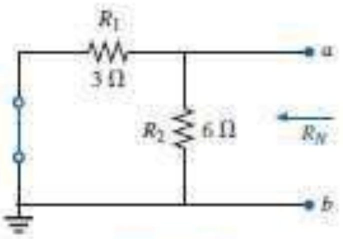


FIG. 9.69
Determining R_N for the network in Fig. 9.68.

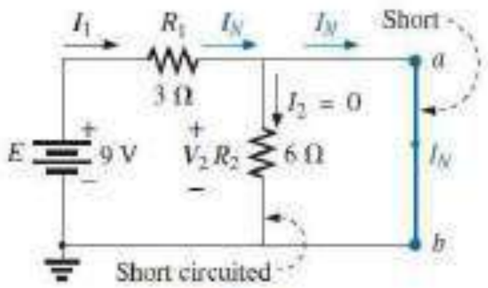


FIG. 9.70
Determining I_N for the network in Fig. 9.68.

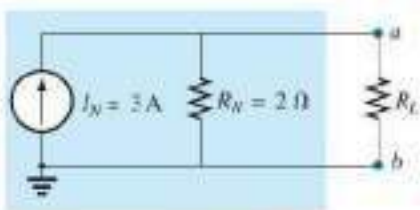


FIG. 9.71
Substituting the Norton equivalent circuit for the network external to the resistor R_L in Fig. 9.67.

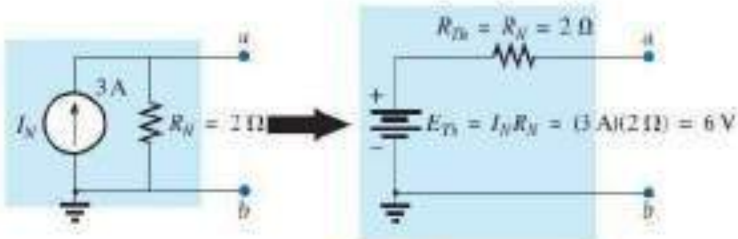


FIG. 9.72
Converting the Norton equivalent circuit in Fig. 9.71 to a Thévenin equivalent circuit.

EXAMPLE 9.13 Find the Norton equivalent circuit for the network external to the 9 _ resistor in Fig. 9.73.

Solution:

Steps 1 and 2: See Fig. 9.74.

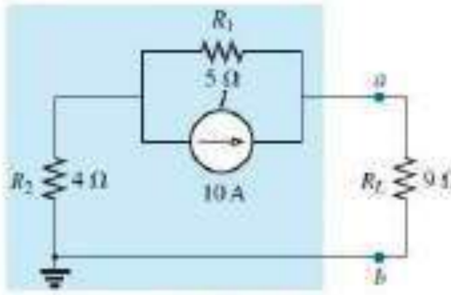


FIG. 9.73
Example 9.13.

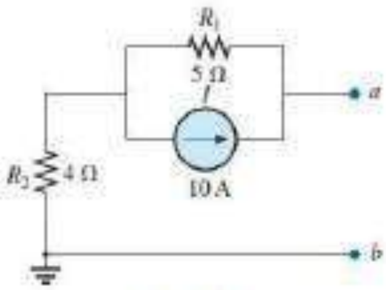


FIG. 9.74
Identifying the terminals of particular interest for the network in Fig. 9.73.

Step 3: See Fig. 9.75, and

$$R_N = R_1 + R_2 = 5\ \Omega + 4\ \Omega = \mathbf{9\ \Omega}$$

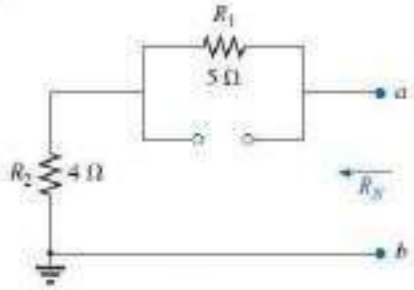


FIG. 9.75
Determining R_N for the network in Fig. 9.74.

Step 4: As shown in Fig. 9.76, the Norton current is the same as the current through the 4 Ω resistor. Applying the current divider rule gives

$$I_N = \frac{R_1 I}{R_1 + R_2} = \frac{(5\ \Omega)(10\ \text{A})}{5\ \Omega + 4\ \Omega} = \frac{50\ \text{A}}{9} = \mathbf{5.56\ A}$$

Step 5: See Fig. 9.77.

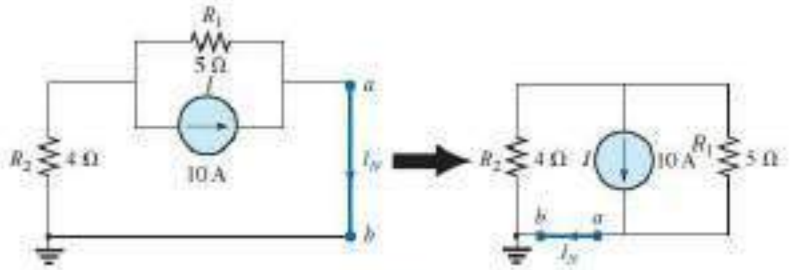


FIG. 9.76
Determining I_N for the network in Fig. 9.74.

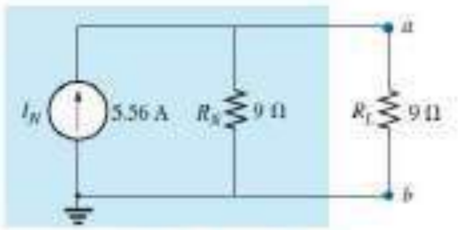


FIG. 9.77
Substituting the Norton equivalent circuit for the network external to the resistor R_L in Fig. 9.73.

EXAMPLE 9.14 (Two sources) Find the Norton equivalent circuit for the portion of the network to the left of *a-b* in Fig. 9.78.

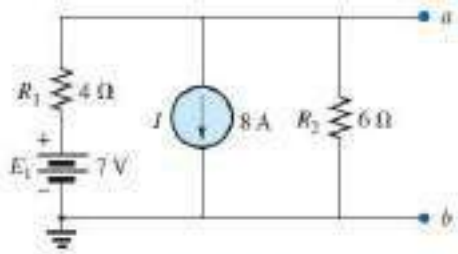


FIG. 9.79
Identifying the terminals of particular interest for the network in Fig. 9.78.

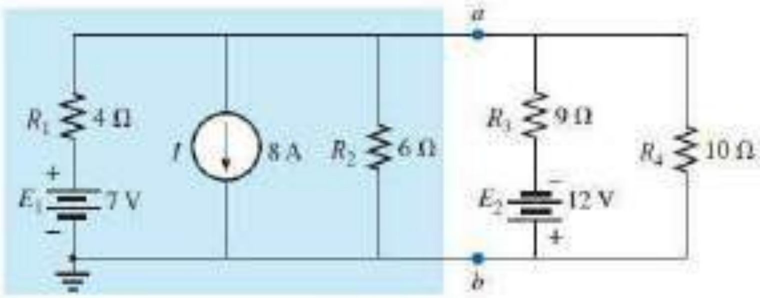


FIG. 9.78
Example 9.14.

Solution:

Steps 1 and 2: See Fig. 9.79.

Step 3: See Fig. 9.80, and

$$R_N = R_1 \| R_2 = 4\ \Omega \| 6\ \Omega = \frac{(4\ \Omega)(6\ \Omega)}{4\ \Omega + 6\ \Omega} = \frac{24\ \Omega}{10} = \mathbf{2.4\ \Omega}$$

Step 4: (Using superposition) For the 7 V battery (Fig. 9.81),

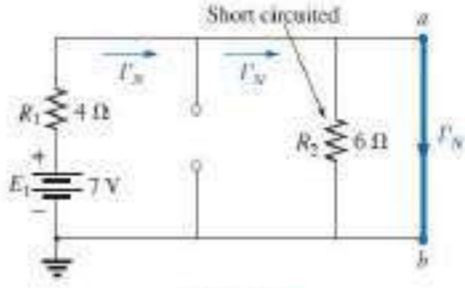


FIG. 9.81
Determining the contribution to I_N from the voltage source E_1.

$$I'_N = \frac{E_1}{R_1} = \frac{7\text{ V}}{4\ \Omega} = 1.75\text{ A}$$

For the 8 A source (Fig. 9.82), we find that both R_1 and R_2 have been "short circuited" by the direct connection between a and b, and

$$I''_N = I = 8\text{ A}$$

The result is

$$I_N = I''_N - I'_N = 8\text{ A} - 1.75\text{ A} = 6.25\text{ A}$$

Step 5: See Fig. 9.83.

9.5 MAXIMUM POWER TRANSFER

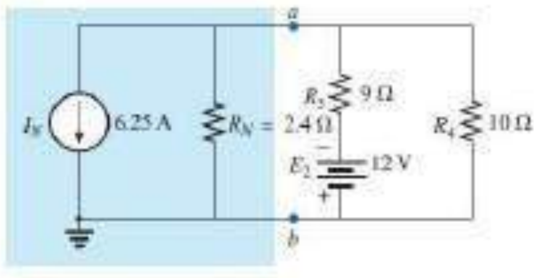


FIG. 9.83
Substituting the Norton equivalent circuit for the network to the left of terminals a-b in Fig. 9.78.

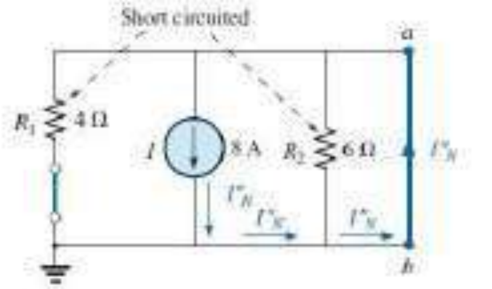


FIG. 9.82
Determining the contribution to I_N from the current source I.

THEOREM

When designing a circuit, it is often important to be able to answer one of the following questions:

What load should be applied to a system to ensure that the load is receiving maximum power from the system?

Conversely:

For a particular load, what conditions should be imposed on the source to ensure that it will deliver the maximum power available?

Even if a load cannot be set at the value that would result in maximum power transfer, it is often helpful to have some idea of the value that will draw maximum power so that you can compare it to the load at hand. For instance, if a design calls for a load of 100 Ω, to ensure that the load receives maximum power, using a resistor of 1 Ω or 1 k Ω results in a power transfer that is much less than the maximum possible.

However, using a load of 82 Ω or 120 Ω probably results in a fairly good level of power transfer. Fortunately, the process of finding the load that will receive maximum power from a particular system is quite straightforward due to the **maximum power transfer theorem,** which states the following:

A load will receive maximum power from a network when its resistance is exactly equal to the Thévenin resistance of the network applied to the load. That is,

$$\boxed{R_L = R_{Th}} \qquad \textbf{(9.2)}$$

In other words, for the Thévenin equivalent circuit in Fig. 9.84, when the load is set equal to the Thévenin resistance, the load will receive maximum power from the network.

Using Fig. 9.84 with $R_L = R_{Th}$, we can determine the maximum power delivered to the load by first finding the current:

$$I_L = \frac{E_{Th}}{R_{Th} + R_L} = \frac{E_{Th}}{R_{Th} + R_{Th}} = \frac{E_{Th}}{2R_{Th}}$$

Then we substitute into the power equation:

$$P_L = I_L^2 R_L = \left(\frac{E_{Th}}{2R_{Th}}\right)^2 (R_{Th}) = \frac{E_{Th}^2 R_{Th}}{4R_{Th}^2}$$

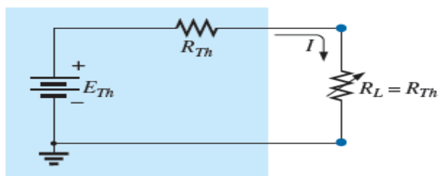


FIG. 9.84
Defining the conditions for maximum power to a load using the Thévenin equivalent circuit.

and

$$\boxed{P_{L_{max}} = \frac{E_{Th}^2}{4R_{Th}}} \qquad (9.3)$$

The total power delivered by a supply such as ETh is absorbed by both the Thévenin equivalent resistance and the load resistance. Any power delivered by the source that does not get to the load is lost to the Thévenin resistance.

Under maximum power conditions, only half the power delivered by the source gets to the load. Now, that sounds disastrous, but remember that we are starting out with a fixed Thévenin voltage and resistance, and the above simply tells us that we must make the two resistance levels equal if we want maximum power to the load. On an efficiency basis, we are working at only a 50% level, but we are content because *we are getting maximum power out of our system*.

The dc operating efficiency is defined as the ratio of the power delivered to the load (*PL*) to the power delivered by the source (*Ps*). That is,

$$\boxed{\eta\% = \frac{P_L}{P_s} \times 100\%} \qquad (9.4)$$

For the situation where $R_L = R_{Th}$,

$$\eta\% = \frac{I_L^2 R_L}{I_L^2 R_T} \times 100\% = \frac{R_L}{R_T} \times 100\% = \frac{R_{Th}}{R_{Th} + R_{Th}} \times 100\%$$

$$= \frac{R_{Th}}{2R_{Th}} \times 100\% = \frac{1}{2} \times 100\% = \mathbf{50\%}$$

If efficiency is the overriding factor, then the load should be much larger than the internal resistance of the supply. If maximum power transfer is desired and efficiency less of a concern, then the conditions dictated by the maximum power transfer theorem should be applied.

A relatively low efficiency of 50% can be tolerated in situations where power levels are relatively low, such as in a wide variety of electronic systems, where maximum power transfer for the given system is usually more important. However, when large power levels are involved, such as at generating plants, efficiencies of 50% cannot be tolerated. In fact, a great deal of expense and research is dedicated to raising power generating and transmission efficiencies a few percentage points. Raising an efficiency level of a 10 MkW power plant from 94% to 95% (a 1% increase) can save 0.1 MkW, or 100 million watts, of power—an enormous saving. In all of the above discussions, the effect of changing the load was discussed for a fixed Thévenin resistance. Looking at the situation from a different viewpoint, we can say

if the load resistance is fixed and does not match the applied Thévenin equivalent resistance, then some effort should be made (if possible) to redesign the system so that the Thévenin equivalent resistance is closer to the fixed applied load.

In other words, if a designer faces a situation where the load resistance is fixed, he or she should investigate whether the supply section should be replaced or redesigned to create a closer match of resistance levels to produce higher levels of power to the load.

For the Norton equivalent circuit in Fig. 9.90, maximum power will be delivered to the load when ,

$$R_L = R_N \qquad (9.5)$$

This result [Eq. (9.5)] will be used to its fullest advantage in the analysis of transistor networks, where the most frequently applied transistor circuit model uses a current source rather than a voltage source.

For the Norton circuit in Fig. 9.90,

$$\boxed{P_{L_{max}} = \frac{I_N^2 R_N}{4}} \quad \text{(W)} \qquad (9.6$$

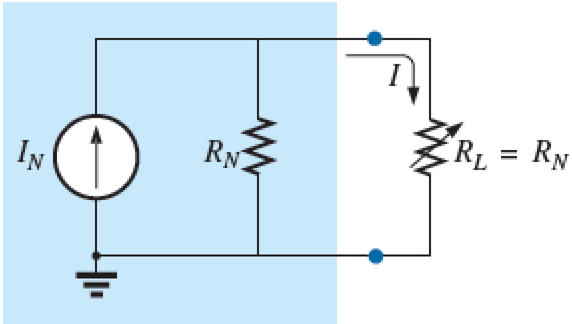


FIG. 9.90

Defining the conditions for maximum power to a load using the Norton equivalent circuit.

To demonstrate that maximum power is indeed transferred to the load under the conditions defined above, consider the Thévenin equivalent circuit in Fig. 9.85. Before getting into detail, however, if you were to guess what value of RL would result in maximum power transfer to RL, you might think that the smaller the value of RL, the better it is because the current reaches a maximum when it is squared in the power equation. The problem is, however, that in the equation $PL = I$ 2 LRL, the load resistance is a multiplier. As it gets smaller, it forms a smaller product. Then again, you might suggest larger values of RL because the output voltage increases, and power is determined by $PL = V$ 2 L /RL. This time, however, the load resistance is in the denominator of the equation and causes the resulting power to decrease. A balance must obviously be made between the load resistance and the resulting current or voltage. The following discussion shows that

maximum power transfer occurs when the load voltage and current are one-half their maximum possible values.

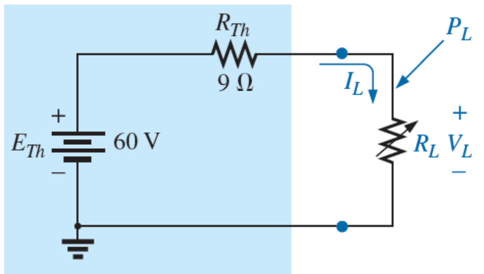


FIG. 9.85

Thévenin equivalent network to be used to validate the maximum power transfer theorem.

For the circuit in Fig. 9.85, the current through the load is determined by

$$I_L = \frac{E_{Th}}{R_{Th} + R_L} = \frac{60\text{ V}}{9\ \Omega + R_L}$$

The voltage is determined by

$$V_L = \frac{R_L E_{Th}}{R_L + R_{Th}} = \frac{R_L(60\text{ V})}{R_L + R_{Th}}$$

and the power by

$$P_L = I_L^2 R_L = \left(\frac{60\text{ V}}{9\ \Omega + R_L}\right)^2 (R_L) = \frac{3600 R_L}{(9\ \Omega + R_L)^2}$$

If we tabulate the three quantities versus a range of values for R_L from 0.1 Ω to 30 Ω, we obtain the results appearing in Table 9.1. Note in particular that when R_L is equal to the Thévenin resistance of 9 Ω, the

power has a maximum value of 100 W, the current is 3.33 A, or one-half its maximum value of 6.67 A (as would result with a short circuit across the output terminals), and the voltage across the load is 30 V, or one-half
its maximum value of 60 V (as would result with an open circuit across its output terminals). As you can see, there is no question that maximum power is transferred to the load when the load equals the Thévenin value.
The power to the load versus the range of resistor values is provided in Fig. 9.86. Note in particular that for values of load resistance less than the Thévenin value, the change is dramatic as it approaches the peak value. However, for values greater than the Thévenin value, the drop is a great deal more gradual. This is important because it tells us the following:
If the load applied is less than the Thévenin resistance, the power to the load will drop off rapidly as it gets smaller. However, if the applied load is greater than the Thévenin resistance, the power to the load will not drop off as rapidly as it increases.

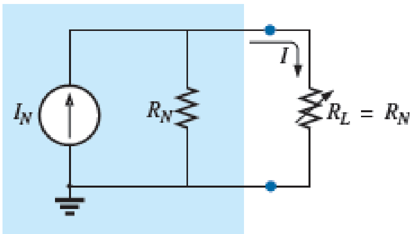


FIG. 9.90
Defining the conditions for maximum power to a load using the Norton equivalent circuit.

In all of the above discussions, the effect of changing the load was discussed for a fixed Thévenin resistance. Looking at the situation from a different viewpoint, we can say

if the load resistance is fixed and does not match the applied Thévenin equivalent resistance, then some effort should be made (if possible) to redesign the system so that the Thévenin equivalent resistance is closer to the fixed applied load.

In other words, if a designer faces a situation where the load resistance is fixed, he or she should investigate whether the supply section should be replaced or redesigned to create a closer match of resistance levels to produce higher levels of power to the load.

For the Norton equivalent circuit in Fig. 9.90, maximum power will be delivered to the load when

$$\boxed{R_L = R_N} \qquad (9.5)$$

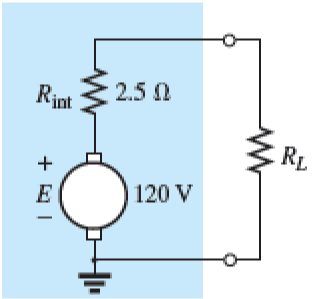
This result [Eq. (9.5)] will be used to its fullest advantage in the analysis of transistor networks, where the most frequently applied transistor circuit model uses a current source rather than a voltage source.

For the Norton circuit in Fig. 9.90,

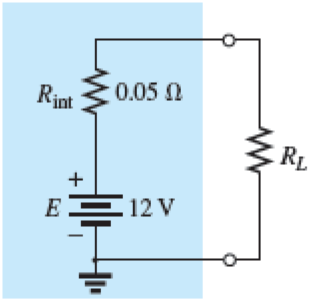
$$\boxed{P_{L_{max}} = \frac{I_N^2 R_N}{4}} \quad (W) \qquad (9.6)$$

EXAMPLE 9.15 A dc generator, battery, and laboratory supply are connected to resistive load R_L in Fig. 9.91.

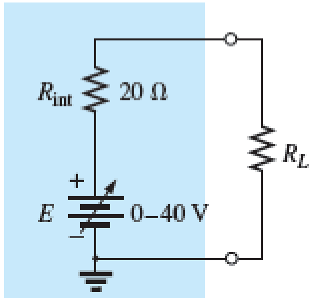
a. For each, determine the value of R_L for maximum power transfer to R_L.
b. Under maximum power conditions, what are the current level and the power to the load for each configuration?
c. What is the efficiency of operation for each supply in part (b)?
d. If a load of 1 kΩ were applied to the laboratory supply, what would the power delivered to the load be? Compare your answer to the level of part (b). What is the level of efficiency?
e. For each supply, determine the value of R_L for 75% efficiency.



(a) dc generator



(b) Battery



(c) Laboratory supply

FIG. 9.91
Example 9.15.

Solutions:

a. For the dc generator,

$$R_L = R_{Th} = R_{int} = \mathbf{2.5\ \Omega}$$

For the 12 V car battery,

$$R_L = R_{Th} = R_{\text{int}} = \mathbf{0.05\ \Omega}$$

For the dc laboratory supply,

$$R_L = R_{Th} = R_{\text{int}} = \mathbf{20\ \Omega}$$

b. For the dc generator,

$$P_{L_{\text{max}}} = \frac{E_{Th}^2}{4R_{Th}} = \frac{E^2}{4R_{\text{int}}} = \frac{(120\ \text{V})^2}{4(2.5\ \Omega)} = \mathbf{1.44\ kW}$$

For the 12 V car battery,

$$P_{L_{\text{max}}} = \frac{E_{Th}^2}{4R_{Th}} = \frac{E^2}{4R_{\text{int}}} = \frac{(12\ \text{V})^2}{4(0.05\ \Omega)} = \mathbf{720\ W}$$

For the dc laboratory supply,

$$P_{L_{\text{max}}} = \frac{E_{Th}^2}{4R_{Th}} = \frac{E^2}{4R_{\text{int}}} = \frac{(40\ \text{V})^2}{4(20\ \Omega)} = \mathbf{20\ W}$$

c. They are all operating under a 50% efficiency level because $R_L = R_{Th}$.

d. The power to the load is determined as follows:

$$I_L = \frac{E}{R_{\text{int}} + R_L} = \frac{40\ \text{V}}{20\ \Omega + 1000\ \Omega} = \frac{40\ \text{V}}{1020\ \Omega} = 39.22\ \text{mA}$$

and $$P_L = I_L^2 R_L = (39.22\ \text{mA})^2(1000\ \Omega) = \mathbf{1.54\ W}$$

The power level is significantly less than the 20 W achieved in part (b). The efficiency level is

$$\eta\% = \frac{P_L}{P_s} \times 100\% = \frac{1.54\ \text{W}}{EI_s} \times 100\% = \frac{1.54\ \text{W}}{(40\ \text{V})(39.22\ \text{mA})} \times 100\%$$
$$= \frac{1.54\ \text{W}}{1.57\ \text{W}} \times 100\% = \mathbf{98.09\%}$$

which is markedly higher than achieved under maximum power conditions—albeit at the expense of the power level.

e. For the dc generator,

$$\eta = \frac{P_o}{P_s} = \frac{R_L}{R_{Th} + R_L} \quad (\eta \text{ in decimal form})$$

and
$$\eta = \frac{R_L}{R_{Th} + R_L}$$

$$\eta(R_{Th} + R_L) = R_L$$
$$\eta R_{Th} + \eta R_L = R_L$$
$$R_L(1 - \eta) = \eta R_{Th}$$

and
$$\boxed{R_L = \frac{\eta R_{Th}}{1 - \eta}} \qquad \textbf{(9.7)}$$

$$R_L = \frac{0.75(2.5\ \Omega)}{1 - 0.75} = \mathbf{7.5\ \Omega}$$

For the battery,

$$R_L = \frac{0.75(0.05\ \Omega)}{1 - 0.75} = \mathbf{0.15\ \Omega}$$

For the laboratory supply,

$$R_L = \frac{0.75(20\ \Omega)}{1 - 0.75} = \mathbf{60\ \Omega}$$

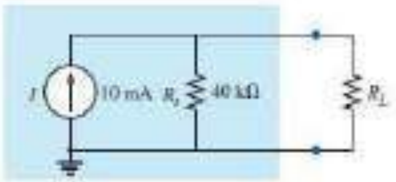


FIG. 9.92
Example 9.16.

EXAMPLE 9.16 The analysis of a transistor network resulted in the reduced equivalent in Fig. 9.92.

a. Find the load resistance that will result in maximum power transfer to the load, and find the maximum power delivered.
b. If the load were changed to 68 kΩ, would you expect a fairly high level of power transfer to the load based on the results of part (a)? What would the new power level be? Is your initial assumption verified?
c. If the load were changed to 8.2 kΩ, would you expect a fairly high level of power transfer to the load based on the results of part (a)? What would the new power level be? Is your initial assumption verified?

Solutions:

a. Replacing the current source by an open-circuit equivalent results in

$$R_{Th} = R_s = 40\ \text{k}\Omega$$

Restoring the current source and finding the open-circuit voltage at the output terminals results in

$$E_{Th} = V_{oc} = IR_s = (10\ \text{mA})(40\ \text{k}\Omega) = 400\ \text{V}$$

For maximum power transfer to the load,

$$R_L = R_{Th} = 40\ \text{k}\Omega$$

with a maximum power level of

$$P_{L_{max}} = \frac{E_{Th}^2}{4R_{Th}} = \frac{(400\ \text{V})^2}{4(40\ \text{k}\Omega)} = 1\ \text{W}$$

b. Yes, because the 68 kΩ load is greater (note Fig. 9.86) than the 40 kΩ load, but relatively close in magnitude.

$$I_L = \frac{E_{Th}}{R_{Th} + R_L} = \frac{400\text{ V}}{40\text{ k}\Omega + 68\text{ k}\Omega} = \frac{400}{108\text{ k}\Omega} \cong 3.7\text{ mA}$$

$$P_L = I_L^2 R_L = (3.7\text{ mA})^2(68\text{ k}\Omega) \cong \mathbf{0.93\ W}$$

Yes, the power level of 0.93 W compared to the 1 W level of part (a) verifies the assumption.

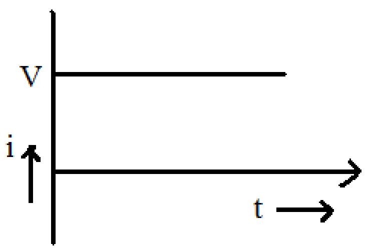
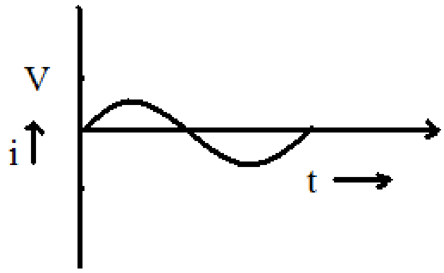
c. No, 8.2 kΩ is quite a bit less (note Fig. 9.86) than the 40 kΩ value.

$$I_L = \frac{E_{Th}}{R_{Th} + R_L} = \frac{400\text{ V}}{40\text{ k}\Omega + 8.2\text{ k}\Omega} = \frac{400\text{ V}}{48.2\text{ k}\Omega} \cong 8.3\text{ mA}$$

$$P_L = I_L^2 R_L = (8.3\text{ mA})^2(8.2\text{ k}\Omega) \cong \mathbf{0.57\ W}$$

Yes, the power level of 0.57 W compared to the 1 W level of part (a) verifies the assumption.

CHAPTER-05
AC CIRCUIT AND RESONANCE

Direct Current		**Alternating Current**	
			
(1)	D.C. always flow in one direction and whose magnitude remains constant.	(1)	A.C. is one which reverse periodically in direction and whose magnitude undergoes a definite cycle changes in definite intervals of time.
(2)	High cost of production.	(2)	Low cost of production
(3)	It is not possible by D.C. Because D.C. is dangerous to the transformer.	(3)	By using transformer A.C. voltage can be decreased or increased.
(4)	Its transmission cost is too high.	(4)	A.C. can be transmitted to a long distance economically.

Definition of A.C. terms :-

Cycle : It is one complete set of +ve and –ve values of alternating quality spread over 360° or 2∏ radan.

Time Period : It is defined as the time required to complete one cycle.

Frequency : It is defined as the reciprocal of time period. i.e. f = 1/ T

Or

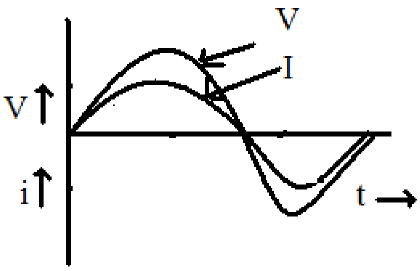
It is defined as the number of cycles completed per second.

Amplitude : It is defined as the maximum value of either +ve half cycle or –ve half cycle.

Phase : It is defined as the angular displacement between two haves is zero.

OR

Two alternating quantity are in phase when each pass through their zero value at the same instant and also attain their maximum value at the same instant in a given cycle.



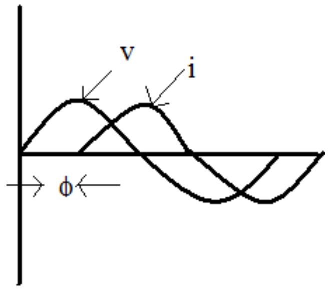
$V = V_m \sin wt$

$i = I_m \sin wt$

Phase Difference :- It is defined as the angular displacement between two alternating quantities.

OR

If the angular displacement between two waves are not zero, then that is known as phase difference. i.e. at a particular time they attain unequal distance.



OR

Two quantities are out of phase if they reach their maximum value or minimum value at different times but always have an equal phase angle between them.

Here $V = V_m \sin wt$

$i = I_m \sin (wt\text{-}\phi)$

In this case current lags voltage by an angle 'ϕ'.

Phasor Diagram :

Generation of Alternating emf :-

Consider a rectangular coil of 'N" turns, area of cross-section is 'A' nt^2 is placed in x-axis in an uniform magnetic field of maximum flux density *Bm web/nt*^{2}. The coil is rotating in the magnetic field with a velocity of w radian / second. At time t = 0, the coil is in x-axis. After interval of time 'dt' second the coil make rotating in anti-clockwise direction and makes an angle 'θ' with x-direction. The perpendicular component of the magnetic field is $\phi = \phi n \cos wt$

According to Faraday's Laws of electro-magnetic Induction

$$e = -N\frac{d\phi}{dt}$$

$$= -N\frac{d}{dt}(\phi_m \cos wt)$$

$$= -N(-\phi_m w \cos wt)$$

$$= Nw\phi_m \sin wt$$

$$= 2\pi f N\phi_m \sin wt (\because w = 2\pi f)$$

$$= 2\pi f N B_m A \sin wt$$

$$e = E_m \sin wt$$

Where $E_m = 2\pi f N B_m A$

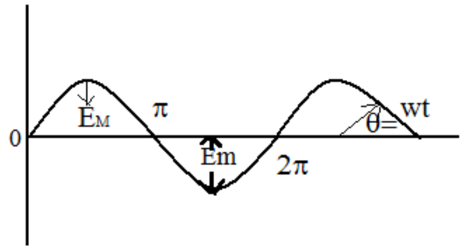
f $\rightarrow$ frequency in Hz

$B_m \rightarrow$ Maximum flux density in Wb/mt^2

Now when θ or wt = 90°

$e = E_m$

i.e. $E_m = 2\pi f N B_m A$



Root Mean Square (R.M.S) Value :$\rightarrow$

The r.m.s. value of an a.c. is defined by that steady (d.c.) current which when flowing through a given circuit for a given time produces same heat as produced by the alternating current when flowing through the same circuit for the same time.

Sinuscdial alternating current is

$i = I_m \sin wt = I_m \sin \theta$

The mean of squares of the instantaneous values of current over one complete cycle

$$= \int_0^{2\pi} \frac{i^2.d\theta}{(2\pi - 0)}$$

The square root of this value is

$$= \sqrt{\int_0^{2\pi} \frac{i^2.d\theta}{2\pi}}$$

$$= \sqrt{\int_0^{2\pi} \frac{(I_m \sin\theta)^2}{2\pi} d\theta}$$

$$= \sqrt{\frac{I_m^2}{2\pi}\int_0^{2\pi} \sin^2\theta . d\theta}$$

$$= \sqrt{\frac{I_m^2}{2\pi}\int_0^{2\pi}\left(\frac{1-\cos 2\theta}{2}\right)d\theta}$$

$$= \sqrt{\frac{I_m^2}{4\pi}\int_0^{2\pi}(1-\cos 2\theta)d\theta}$$

$$= \sqrt{\frac{I_m^2}{4\pi}\left[\frac{\theta - \sin 2\theta}{2}\right]_0^{2\pi}}$$

$$= \sqrt{\frac{I_m^2}{4\pi}\int_0^{2\pi}\left(2\pi - \frac{\sin 4\pi}{2}\right)d\theta}$$

$$= \sqrt{\frac{I_m^2}{4\pi}\int_0^{2\pi}(2\pi - 0)}$$

$$= \sqrt{\frac{I_m^2}{2}} = \frac{I_m}{\sqrt{2}}$$

$$I_{r.m.s} = \frac{I_m}{\sqrt{2}} = 0.707\ I_m$$

Average Value :→

The average value of an alternating current is expressed by that steady current (d.c.) which transfers across any circuit the same charge as it transferred by that alternating current during the sae time.

The equation of the alternating current is $i = I_m \sin\theta$

$$I_{av} = \int_0^{\pi} \frac{i\ .d\theta}{(\pi - 0)}$$

$$= \int_0^{\pi} \frac{I_m.\sin\theta}{\pi} d\theta \quad = \frac{I_m}{\pi}\int_0^{\pi} \sin\theta . d\theta$$

$$= \frac{I_m}{\pi}\left[-\cos\theta\right]_0^{\pi} \quad = \frac{I_m}{\pi}\left[-\cos\pi - (\cos 0^0\right]$$

$$= \frac{I_m}{\pi}\left[1 - 0(-1)\right]$$

$$I_{av} = \frac{2I_m}{\pi}$$

$$I_{av} = \frac{2 \times Maximum\ Current}{\pi}$$

Hence, $I_{av} = 0.637 I_m$

The average value over a complete cycle is zero

Amplitude factor/ Peak factor/ Crest factor :- It is defined as the ratio of maximum value to r.m.s value.

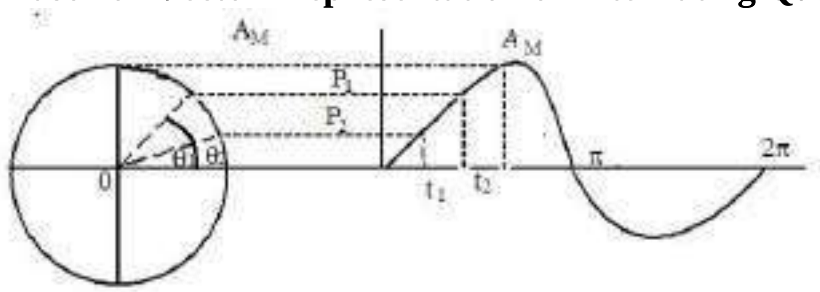
$$Ka = \frac{Maximum Value}{R.M.S.Value} = \frac{I_m}{\frac{I_m}{\sqrt{2}}} = \sqrt{2} = 1.414$$

Form factor : - It is defined as the ratio of r.m.s value to average value.

$$Kf = \frac{r.m.s.Value}{Average.Value} = \frac{0.707 I_m}{0.637 I_m} = \sqrt{2} = 1.414$$

$$Kf = 1.11$$

Phasor or Vector Representation of Alternating Quantity :→



An alternating current or voltage, (quantity) in a vector quantity which has magnitude as well as direction. Let the alternating value of current be represented by the equation e = E_m Sin wt. The projection of E_m on Y-axis at any instant gives the instantaneous value of alternating current. Since the instantaneous values are continuously changing, so they are represented by a rotating vector or phasor. A phasor is a vector rotating at a constant angular velocity

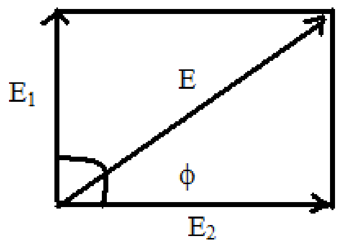
At $t_1, e_1 = E_{m_1} \sin wt_1$

At $t_2, e_2 = E_{m_2} \sin wt_2$

Addition of two alternating Current :→

Let $e_1 = E_{m_1} \sin wt$

$e_2 = E_{m_2} \sin(wt - \phi)$



The sum of two sine waves of the same frequency is another sine wave of same frequency but of a different maximum value and Phase.

$$e = \sqrt{e_1^2 + e_2^2 + 2e_1 e_2 \cos\phi}$$

Phasor Algebra :→

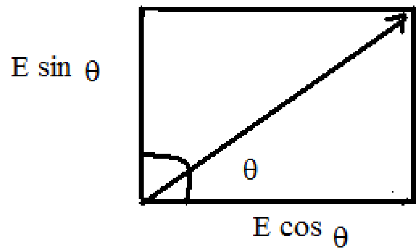
A vector quantity can be expressed in terms of

(i) Rectangular or Cartesian form

(ii) Trigonometric form

(iii) Exponential form

(iv) Polar form



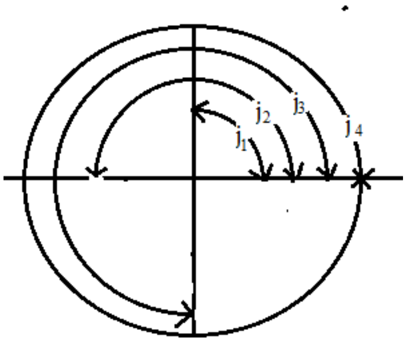
$$E = a + jb$$
$$= E(\cos\theta + j\sin\theta)$$

Where a = E cos θ is the active part

b = E sin θ is the reactive part

$\theta = \tan^{-1}\left(\frac{b}{a}\right) =$ Phase angle

$$j = \sqrt{-1}(90^o)$$
$$j^2 = -1(180^o)$$
$$j^3 = -j(270^o)$$
$$j^4 = 1 \quad (360^o)$$



(i) Rectangular for :-

$$E = a \pm jb$$
$$\tan\theta = b/a$$

(ii) Trigonometric form :-

$$E = E(\cos\theta \pm j\sin\theta)$$

(iii) Exponential form :-

$$E = Ee^{\pm j\theta}$$

(iv) Polar form :-

$$E = E/\pm e \qquad (E = \sqrt{a^2 + b^2})$$

Addition or Subtration :-

$$E_1 = a_1 + jb_1$$
$$E_2 = a_2 + jb_2$$
$$E_1 \pm E_2 = (a_1 + a_2) \pm (b_1 + b_2$$
$$\phi = \tan^{-1}\left(\frac{b_1 + b_2}{a_1 + a_2}\right)$$

Multiplication : -

$$E_1 \times E_2 = (a_1 + ja_1) \pm (a_1 + jb_2)$$
$$= (a_1a_2 - b_1b_2) + j(a_1a_2 + b_1b_2)$$

$$\phi = \tan^{-1}\left(\frac{a_1b_2 + b_1a_2}{a_1a_2 - b_1b_2}\right)$$

$$E_1 = E_1\angle\theta_1$$

$$E_2 = E_2\angle\theta_2$$

$$E_1 \times E_2 = E_1E_2 \quad \angle\phi_1 + \phi_2$$

Division :-

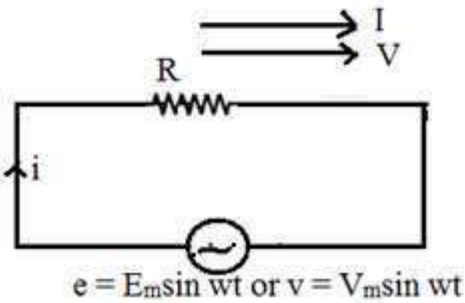
$$E_1 = E_1\angle\theta_1$$

$$E_2 = E_2\angle\theta_2$$

$$\frac{E_1}{E_2} = \frac{E_1\angle\theta_1}{E_2\angle\theta_2} = \frac{E_1}{E_2}\angle\theta_1 - \theta_2$$

A.C. through Pure Resistance :→

Let the resistance of R ohm is connected across to A.C supply of applied voltage



$e = E_m \sin wt$ -------------------------- (1)

Let ‘I’ is the instantaneous current .

Here $e = iR$

$\Rightarrow i = e/R$

$i = E_m\sin wt / R$------------------------ (2)

By comparing equation (1) and equation (2) we get alternating voltage and current in a pure resistive circuit are in phase

Instantaneous power is given by

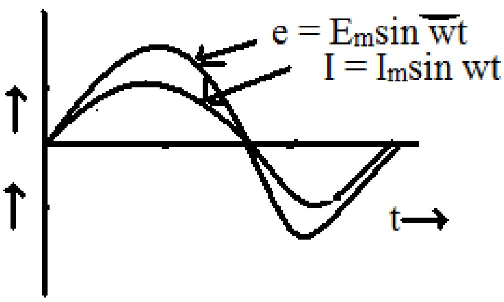
P = ei

$= E_m \sin wt \, . \, I_m \sin wt$

$= E_m I_m \sin^2 wt$

$$= \frac{E_m I_m}{2} . 2\sin^2 wt$$

$$= \frac{E_m}{\sqrt{2}} . \frac{I_m}{\sqrt{2}} . (1 - \cos 2wt)$$



$$P = \frac{E_m}{\sqrt{2}} . \frac{I_m}{\sqrt{2}} - \frac{E_m}{\sqrt{2}} . \frac{I_m}{\sqrt{2}} \cos 2wt$$

i.e. $P = \frac{V_m}{\sqrt{2}} . \frac{I_m}{\sqrt{2}} - \frac{V_m}{\sqrt{2}} . \frac{I_m}{\sqrt{2}} \cos 2wt$

Where $\frac{V_m}{\sqrt{2}}.\frac{I_m}{\sqrt{2}}$ is called constant part of power.

$\frac{V_m}{\sqrt{2}}.\frac{I_m}{\sqrt{2}}.\cos 2wt$ is called fluctuating part of power.

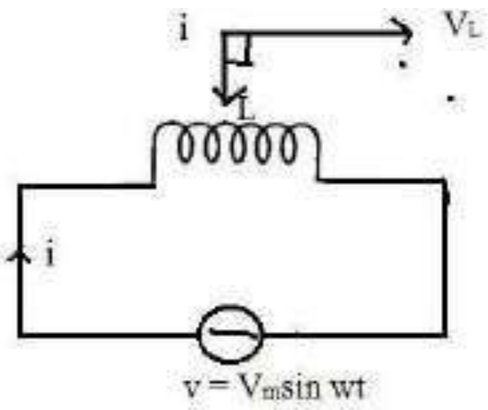
The fluctuating part $\frac{V_m I_m}{2}.\cos 2wt$ of frequency double that of voltage and current waves.

Hence power for the whole cycle is $P = \frac{V_m}{\sqrt{2}}.\frac{I_m}{\sqrt{2}} = V_{rms}.I_{rms}$

$\boxed{\Rightarrow \text{P = VI watts}}$

A.C through Pure Inductance :→

Let inductance of ‘L’ henry is connected across the A.C. supply

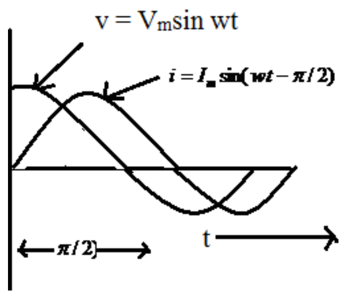


$v = V_m \sin wt$ -------------------------- (1)

According to Faraday’s laws of electromagnetic inductance the emf induced across the inductance

$$V = L\frac{di}{dt}$$

$\frac{di}{dt}$ is the rate of change of current



$V_m \sin wt = L\frac{di}{dt}$

$\frac{di}{dt} = \frac{V_m \sin wt}{L}$

$\Rightarrow di = \frac{V_m}{L}\sin wt.dt$

Integrating both sides,

$\int di = \int \frac{V_m}{L}\sin wt.dt$

$i = \frac{V_m}{L}\left(-\frac{\cos wt}{w}\right)$

$$i = -\frac{V_m \cos wt}{wL}$$

$$i = -\frac{V_m}{wL}\cos wt$$

$$i = -\frac{V_m}{wL}\sin\left(wt - \frac{\pi}{2}\right)$$

$$= -\frac{V_m}{X_L}\sin\left(wt - \frac{\pi}{2}\right) \quad [\because X_L = 2\pi fL = wL]$$

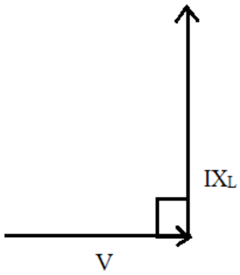
Maximum value of i is $I_m = \frac{V_m}{X_L}$ when $\sin\left(wt - \frac{\pi}{2}\right)$ is unity.

Hence the equation of current becomes $i = I_m \sin(wt - \pi/2)$

So we find that if applied voltage is rep[resented by $v = V_m \sin wt$, then current flowing in a purely inductive circuit is given by

$$i = I_m \sin(wt - \pi/2)$$

Here current lags voltage by an angle $\pi/2$ Radian.

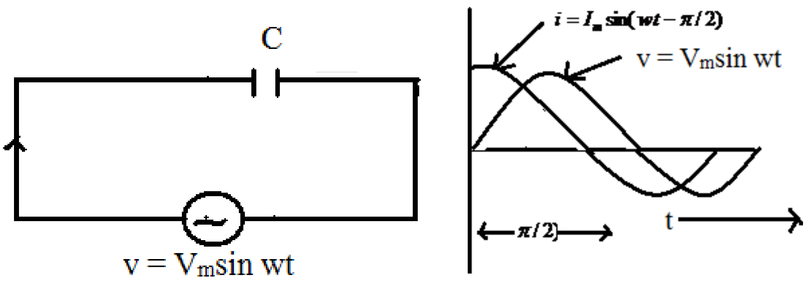


Power factor $= \cos\phi$
$= \cos 90°$
$= 0$

Power Consumed $= VI\cos\phi$
$= VI \times 0$
$= 0$

Hence, the power consumed by a purely Inductive circuit is zero.

A.C. Through Pure Capacitance : →



Let a capacitance of ‘C” farad is connected across the A.C. supply of applied voltage

$$v = V_m \sin wt \text{-----------------------------------} (1)$$

Let ‘q’ = change on plates when p.d. between two plates of capacitor is ‘v’

$q = cv$

$q = cV_m \sin wt$

$$\frac{dq}{dt} = c\frac{d}{dt}(V_m \sin wt)$$

$$i = cV_m \sin wt$$

$$= wcV_m \cos wt$$

$$= \frac{V_m}{1/wc} = \cos wt$$

$$= \frac{V_m}{Xc} = \cos wt$$ [$\because X_c = \frac{1}{wc} = \frac{1}{2\pi fc}$ is known as capacitive reactance in ohm.]

$$= I_m \cos wt$$

$$= I_m \sin(wt + \pi/2)$$

Here current leads the supply voltage by an angle π/2 radian.

Power factor = cos ϕ

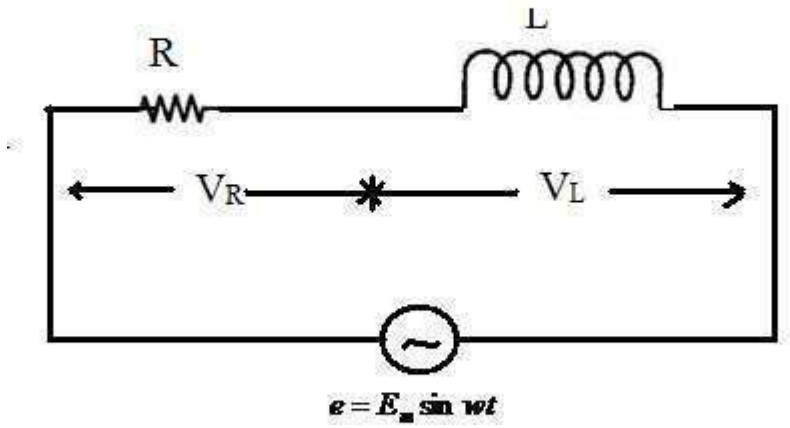
= cos 90° = 0

Power Consumed = VI cos ϕ

= VI × 0 = 0

The power consumed by a pure capacitive circuit is zero.

A.C. Through R-L Series Circuit : →



The resistance of R-ohm and inductance of L-henry are connected in series across the A.C. supply of applied voltage

$e = E_m \sin wt$ -----------------------------(1)

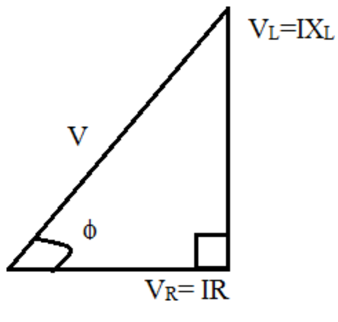
$$V = V_R + jV_L$$

$$= \sqrt{V_R^2 + V_L^2}\angle\phi = \tan^{-1}\left(\frac{X_L}{R}\right)$$

$$= \sqrt{(IR)^2 + (IX_L)^2}\angle\phi = \tan^{-1}\left(\frac{X_L}{R}\right)$$

$$= I\sqrt{R^2 + X_L^2}\angle\phi = \tan^{-1}\left(\frac{X_L}{R}\right)$$

$$V = IZ\angle\phi = \tan^{-1}\left(\frac{X_L}{R}\right)$$



Where $Z = \sqrt{R^2 + X_L{}^2}$

$= R + jX_L$ is known as impedance of R-L series Circuit.

$$I = \frac{V}{Z\angle\phi} = \frac{E_m \sin wt}{Z\angle\phi}$$

$$I = I_m \sin(wt - \phi)$$

Here current lags the supply voltage by an angle ϕ.

Power Factor :$\rightarrow$ It is the cosine of the angle between the voltage and current.

OR

It is the ratio of active power to apparent power.

OR

It is the ratio of resistance to inpedence .

Power :$\rightarrow$

$$= v.i$$

$$= V_m \sin wt . I_m \sin(wt - \phi)$$

$$= V_m I_m \sin wt . \sin(wt - \phi)$$

$$= \frac{1}{2} V_m I_m 2 \sin wt . \sin(wt - \phi)$$

$$= \frac{1}{2} V_m I_m [\cos\phi - \cos 2(wt - \phi)]$$

Obviously the power consists of two parts.

(i) a constant part $\frac{1}{2} V_m I_m \cos\phi$ which contributes to real power.

(ii) a pulsating component $\frac{1}{2} V_m I_m \cos(2wt - \phi)$ which has a frequency twice that of the voltage and current. It does not contribute to actual power since its average value over a complete cycle is zero.

Hence average power consumed

$$= \frac{1}{2} V_m I_m \cos\phi$$

$$= \frac{V_m}{\sqrt{2}} . \frac{I_m}{\sqrt{2}} \cos\phi$$

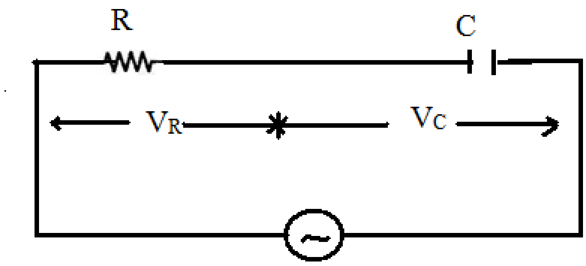
$$= VI \cos\phi$$

Where V & I represents the r.m.s value.

A.C. Through R-C Series Circuit : $\rightarrow$

The resistance of 'R'-ohm and capacitance of 'C' farad is connected across the A.C. supply of applied voltage

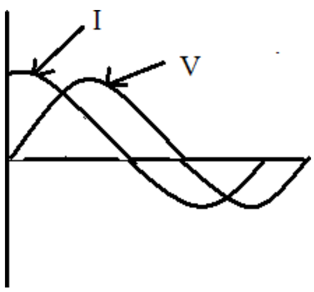
$e = E_m \sin wt$ ------------------------------(1)



$$V = V_R + (-jV_C)$$
$$= IR + (-jIX_C)$$
$$= I(R - jX_C)$$
$$V = IZ$$

Where $Z = R - jX_C = \sqrt{R^2 + X_C{}^2}$ is known as impedance of R-C series Circuit.

$$Z = R - jX_C$$
$$= \sqrt{R^2 + X_C{}^2}$$
$$\angle -\phi = \tan^{-1}\left(\frac{X_C}{R}\right)$$
$$V = IZ\angle -\phi$$
$$\Rightarrow I = \frac{V}{Z\angle -\phi}$$
$$= \frac{E_m \sin wt}{Z\angle -\phi}$$
$$= \frac{E_m}{Z\angle}\sin(wt + \phi)$$

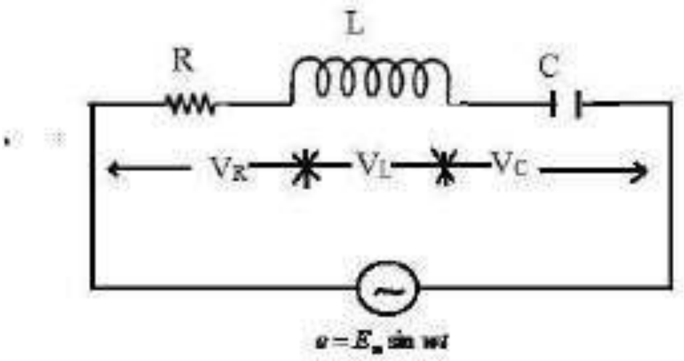


$$\Rightarrow I = I_m \sin(wt + \phi)$$

Here current leads the supply voltage by an angle ‘ϕ’.

A.C. Through R-L-C Series Circuit : →

Let a resistance of ‘R’-ohm inductance of ‘L’ henry and a capacitance of ‘C’ farad are connected across the A.C. supply in series of applied voltage



$e = E_m \sin wt$ --------------------------------- (1)

$$e = \vec{V_R} + \vec{V_L} + \vec{V_C}$$
$$= V_R + jV_L - jV_C$$
$$= V_R + j(V_L - V_C)$$
$$= I_R + j(IX_L - IX_C)$$
$$= I[R + j(X_L - X_C)]$$
$$= I\sqrt{R^2 + (X_L - X_C)^2} \quad \angle \pm \phi = \tan^{-1}\left(\frac{X_L - X_C}{R}\right)$$
$$= IZ\angle \pm \phi$$

Where $Z = I\sqrt{R^2 + (X_L - X_C)^2}$ is known as the impedance of R-L-C Series Circuit.

If $X_L > X_C$, then the angle is +ve.

If $X_L < X_C$, then the angle is -ve.

Impedance is defined as the phasor sum of resistance and net reactance

$$e = IZ\angle \pm \phi$$
$$\Rightarrow I = \frac{e}{Z\angle \pm \phi} IZ\angle \pm \phi \quad = \frac{E_m \sin wt}{Z\angle \pm \phi} \quad = I_m \sin(wt \pm \phi)$$

(1) If $X_L > X_C$, then P.f will be lagging.

(2) If $X_L < X_C$, then, P.f will be leading.

(3) If $X_L = X_C$, then, the circuit will be resistive one. The p.f. becomes unity and the resonance occurs.

REASONANCE

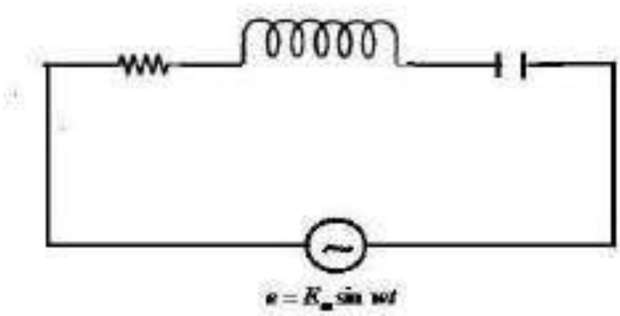
It is defined as the resonance in electrical circuit having passive or active elements represents a particular state when the current and the voltage in the circuit is maximum and minimum with respect to the magnitude of excitation at a particular frequency and the impedances being either minimum or maximum at unity power factor

Resonance are classified into two types.

(1) Series Resonance

(2) Parallel Resonance

(1) Series Resonance :- Let a resistance of 'R' ohm, inductance of 'L' henry and capacitance of 'C' farad are connected in series across A.C. supply



$e = E_m \sin wt$

The impedance of the circuit

$$Z = R + j(X_L - X_C)]$$
$$Z = \sqrt{R^2 + (X_L - X_C)^2}$$

The condition of series resonance:

The resonance will occur when the reactive part of the line current is zero

The p.f. becomes unity.

The net reactance will be zero.

The current becomes maximum.

At resonance net reactance is zero

$$X_L - X_C = 0$$
$$\Rightarrow X_L = X_C$$
$$\Rightarrow W_o L = \frac{1}{W_o C}$$
$$\Rightarrow W_o^2 LC = 1$$
$$\Rightarrow W_o^2 = \frac{1}{LC}$$
$$\Rightarrow W_o = \frac{1}{\sqrt{LC}}$$
$$\Rightarrow 2\pi f_o = \frac{1}{\sqrt{LC}}$$
$$\Rightarrow f_o = \frac{1}{2\pi\sqrt{LC}}$$

Resonant frequency $(f_o) = \frac{1}{2\pi} \cdot \frac{1}{\sqrt{LC}}$

Impedance at Resonance

$Z_0 = R$

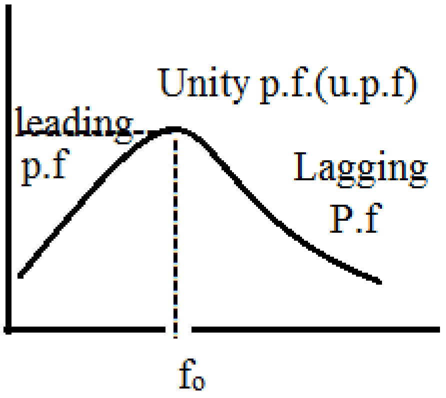
Current at Resonance

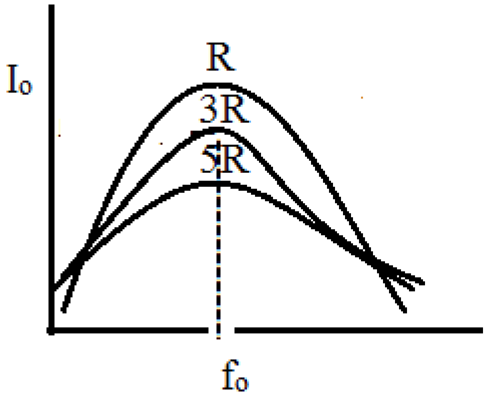
$$I_o = \frac{V}{R}$$

Power factor at resonance

$$p.f. = \frac{R}{Z_o} = \frac{R}{R} = 1 \qquad [\because Z_o = R]$$

Resonance Curve :-





At low frequency the X_c is greater and the circuit behaves leading and at high frequency the X_L becomes high and the circuit behaves lagging circuit.

If the resistance will be low the curve will be stiff (peak).

- If the resistance will go oh increasing the current goes on decreasing and the curve become flat.

Band Width :→

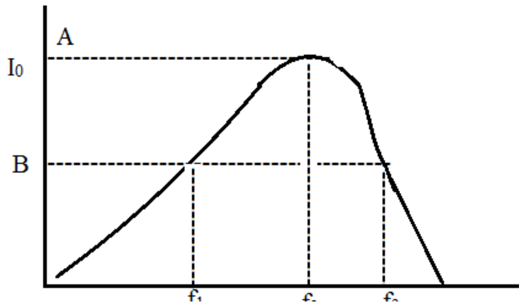
At point 'A' the power loss is I_0^2R.

The frequency is f_0 which is at resonance.

At point 'B' the power loss is $\frac{I_0^2 R}{2}$.

The power loss is 50% of the power loss at point

'A"/



Hence the frequencies corresponding to point 'B' is known as half power frequencies f_1 & f_2.

f_1 = Lower half power frequency

$$f_1 = f_0 - \frac{R}{4\pi L}$$

F_2 = Upper half power frequency

$$f_2 = f_0 + \frac{R}{4\pi L}$$

Band width (B.W.) is defined as the difference between upper half power frequency ad lower half power frequency.

B.W. = $f_2 - f_1 = \frac{R}{2\pi L}$

Selectivity : $\rightarrow$

Selectivity is defined as the ratio of Band width to resonant frequency

$$\text{Selectivity} = \frac{B.W.}{f_0} = \frac{R}{2\pi L} \qquad Selectivity = \frac{R}{2\pi f_o L}$$

Quality Factor (Q-factor) :$\rightarrow$

It is defined as the ratio of $2\pi \times$ Maximum energy stored to energy dissipated per cycle

$$\text{Q-factor} = \frac{2\pi \times \frac{1}{2} LI_0^2}{I^2RT}$$

$$= \frac{\pi L\left(\sqrt{2}I\right)^2}{I^2RT}$$

$$= \frac{\pi L.2I^2}{I^2RT}$$

$$= \frac{\pi L.2I^2}{I^2RT}$$

$$= \frac{2\pi L.}{RT}$$

$$\boxed{\text{Quality factor} = \frac{2\pi f_0 L.}{R}} \qquad \left[\text{Q} \frac{1.}{I} = f_0\right]$$

Quality factor is defined as the reciprocal of power factor.

$$\boxed{\text{Q factor} = \frac{1.}{\cos\phi}}$$

It is the reciprocal of selectivity.

$$\text{Q-factor Or Magnification factor} = \frac{\text{Voltage across Inductor.}}{\text{Voltage across resistor}}$$

$$= \frac{I_0 X_L}{I_0 R}$$

$$= \frac{X_L}{R}$$

$$= \frac{2\pi f_0 L}{R} = \frac{W_0 L}{R}$$

$$\boxed{\text{Q- factor} = \frac{W_0 L}{R}}$$

$$\text{Q-factor factor} = \frac{\text{Voltage across Capacotor.}}{\text{Voltage across resistor}}$$

$$= \frac{I_0 X_c}{I_0 R}$$

$$= \frac{X_C}{R}$$

$$= \frac{1}{2\pi f_0 C} = \frac{1}{2\pi f_0 CR}$$

$$\boxed{\text{Q-factor} = \frac{1}{W_0 CR}}$$

$$Q^2 = \frac{W_0 L}{R} \times \frac{1}{W_0 CR}$$

$$Q^2 = \frac{1}{R^2 C}$$

$$Q = \sqrt{\frac{1}{R^2 C}}$$

$$\boxed{Q = \frac{1}{R}\sqrt{\frac{L}{C}}}$$

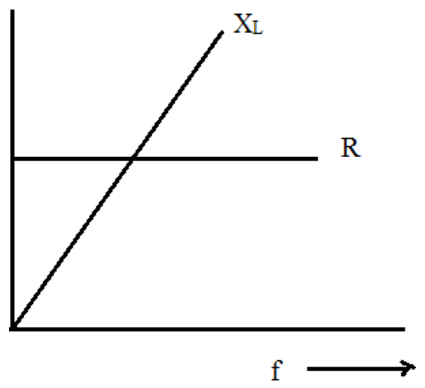
Graphical Method :→

(1) Resistance is independent of frequency It represents a straight line.

(2) Inductive Reactance $X_L = 2\pi fL$

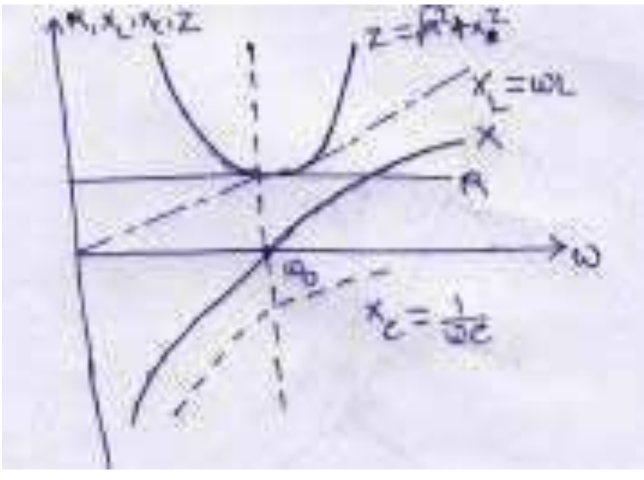
It is directly proportional to frequency. As the frequency increases , X_L increases

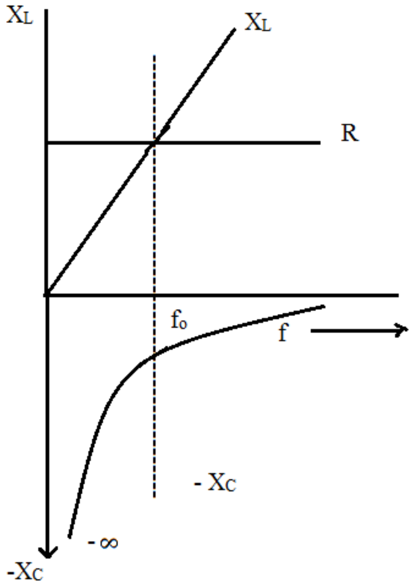
(3) Capacitive Reactance $X_C = = \frac{1}{2\pi fC}$



It is inversely proportional to frequency. As the frequency increases, X_C decreases.

When frequency increases, X_L increases and X_C decreases from the higher value.



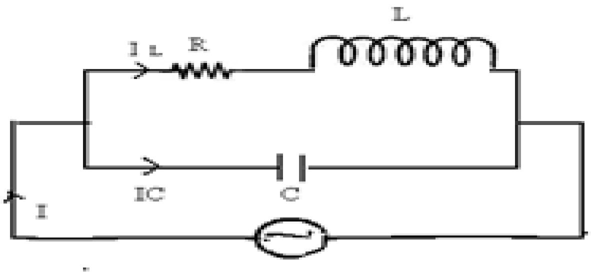


At a certain frequency. $X_L = X_C$

That particular frequency is known as Resonant frequency.

Variation of circuit parameter in series resonance:

(2) Parallel Resonance :- Resonance will occur when the reactive part of the line current is zero.



At resonance,

$I_C - I_L \sin\phi = 0$

$I_C = I_L \sin\phi$

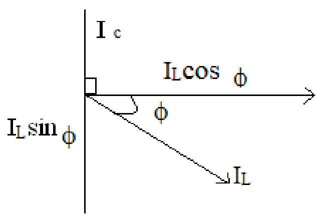
$$\Rightarrow \frac{V}{X_C} = \frac{V}{\sqrt{R^2 + X_L^2}} \sin\phi$$

$$\Rightarrow \frac{V}{X_C} = \frac{V}{\sqrt{R^2 + X_L^2}} \times \frac{X_L}{\sqrt{R^2 + X_L^2}}$$

$$\Rightarrow \frac{1}{X_C} = \frac{X_L}{R^2 + X_L^2}$$

$$\Rightarrow R^2 + X_L^2 = X_L.X_C$$

$$\Rightarrow Z^2 == X_L.X_C = W_0 L \times \frac{1}{W_0 C}$$



$$Z^2 = \frac{L}{C}$$

$$\Rightarrow R^2 + X_L{}^2 = \frac{L}{C}$$

$$\Rightarrow R^2 + (2\pi f_0 L)^2 = \frac{L}{C}$$

$$\Rightarrow R^2 + 4\pi^2 f_0{}^2 L^2 = \frac{L}{C}$$

$$\Rightarrow 4\pi^2 f_0{}^2 L^2 = \frac{L}{C} - R^2$$

$$\Rightarrow f_0{}^2 = \frac{1}{4\pi^2 f_0{}^2 L^2} = \left(\frac{L}{C} - R^2\right)$$

$$\Rightarrow f_0 = \frac{1}{2\pi}\sqrt{\frac{1}{LC} - \frac{R^2}{L^2}}$$

f_0 = Resonant frequency in parallel circuit.

Current at Resonance = $I_L \cos\phi$

$$= \frac{V}{\sqrt{R^2 + X_L{}^2}} \cdot \frac{R}{\sqrt{R^2 + X_L{}^2}}$$

$$= \frac{VR}{R^2 + X_L^2}$$

$$= \frac{VR}{Z^2}$$

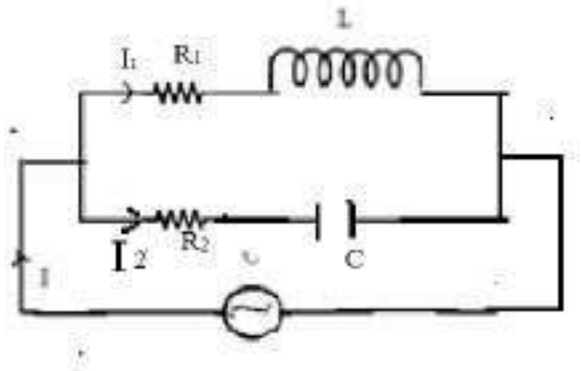
$$= \frac{VR}{L/C} = \frac{V}{L/RC}$$

$$= \frac{\text{V}}{\text{Dynamic Impedence}}$$

$L/RC \rightarrow$ Dynamic Impedance of the circuit.

or, dynamic impedances is defined as the impedance at resonance frequency in parallel circuit.

Parallel Circuit :→



The parallel resonance condition:

When the reactive part of the line current is zero.
The net reactance is zero.
The line current will be minimum.
The power factor will be unity

Impedance $Z_1 = R_1 + jX_L$

$Z_2 = R_2 - jX_C$

Admittance $Y_1 = \frac{1}{Z_1} = \frac{1}{R_1 + jX_L}$

$= \frac{(R_1 + jX_L)}{(R_1 + jX_L)(R_1 - jX_L)}$

$= \frac{R_1 + jX_L}{R_1^2 + X_L^2}$

$Y_1 = \frac{R_1}{R_1^2 + X_L^2} - j\frac{X_L}{R_1^2 + X_L^2}$

Admittance $Y_2 = \frac{1}{Z_2} = \frac{1}{R_1 + jX_C}$

$= \frac{(R_2 + jX_C)}{(R2_1 - jX_C)(R_2 + jX_C)}$

$= \frac{R_2 + jX_L}{R_2^2 + X_C^2}$

$Y_2 = \frac{R_2}{R_2^2 + X_C^2} + j\frac{X_C}{R_2^2 + X_C^2}$

Total Admittance Admittance $\left(\frac{1}{Z}\right) = \frac{1}{Z_1} + \frac{1}{Z_2}$

$\Rightarrow Y = Y_1 + Y_2$

$\Rightarrow Y = \frac{R_1}{R_1^2 + X_L^2} - j\frac{X_L}{R_1^2 + X_L^2} + \frac{R2}{R_2^2 + X_C^2} + j\frac{X_C}{R_2^2 + X_C^2}$

$\Rightarrow Y = \frac{R_1}{R_1^2 + X_L^2} + \frac{R2}{R_2^2 + X_C^2} - j\left(\frac{X_L}{R_1^2 + X_L^2} - \frac{X_C}{R_2^2 + X_C^2}\right)$

At Resonance,

$\frac{X_L}{R_1^2 + X_L^2} - \frac{X_C}{R_2^2 + X_C^2} = 0$

$\Rightarrow \frac{X_L}{R_1^2 + X_L^2} = \frac{X_C}{R_2^2 + X_C^2}$

$\Rightarrow X_L\left(R_2^2 + X_C^2\right) = X_C\left(R_1^2 + X_L^2\right)$

$\Rightarrow 2\pi fL\left(R_2^2 + \frac{1}{4\pi^2 f^2 C^2}\right) = \frac{1}{2\pi fC}\left(R_1^2 + 4\pi^2 f^2 L^2\right)$

$\Rightarrow 2\pi fLR_2^2 + \frac{L}{2\pi fC^2} = \frac{R_1^2}{2\pi fC} + \frac{2\pi fL^2}{C}$

$$\Rightarrow \frac{L}{2\pi f C^2} - \frac{R_1^2}{2\pi f C} = \frac{2\pi f L^2}{C} - 2\pi f L R_2^2$$

$$\Rightarrow \frac{1}{2\pi f C}\left(\frac{L}{C} - R_1^2\right) = 2\pi f L\left(\frac{L}{C} - R_2^2\right)$$

$$\Rightarrow 4\pi^2 f^2 LC = \frac{\frac{L}{C} - R_1^2}{\frac{L}{C} - R_2^2} = \frac{L - CR_1^2}{L - CR_2^2}$$

$$\Rightarrow 4\pi^2 f^2 = \frac{1}{LC}\left(\frac{L - CR_1^2}{L - CR_2^2}\right)$$

$$\Rightarrow f^2 = \frac{1}{4\pi^2 LC}\left(\frac{L - CR_1^2}{L - CR_2^2}\right)$$

$$\Rightarrow f = \frac{1}{2\pi\sqrt{LC}}\sqrt{\left(\frac{L - CR_1^2}{L - CR_2^2}\right)}$$

$$\Rightarrow f = \frac{1}{2\pi}\sqrt{\left(\frac{L - CR_1^2}{L^2C - LC^2R_2^2}\right)}$$

f is called Resonant frequency.

If $R^2 = 0$

Then $f = \frac{1}{2\pi}\sqrt{\frac{L - CR_1^2}{L^2C}}$

$$= \frac{1}{2\pi L}\sqrt{\frac{L - CR_1^2}{C}}$$

$$= \frac{1}{2\pi L}\sqrt{\frac{L}{C} - R_1^2}$$

$$= \frac{1}{2\pi}\sqrt{\frac{L}{L^2C} - \frac{R_1^2}{L^2}}$$

$$\boxed{f = \frac{1}{2\pi}\sqrt{\frac{L}{LC} - \frac{R_1^2}{L^2}}}$$

If R_1 and $R_2 = 0$, then

$$f = \frac{1}{2\pi}\sqrt{\frac{L}{L^2C}}$$

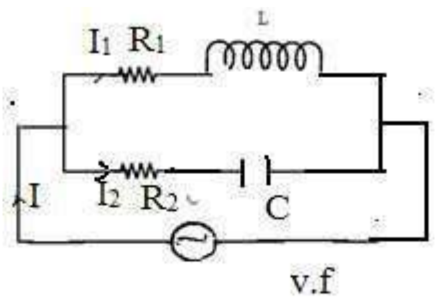
$$\boxed{f = \frac{1}{2\pi}\sqrt{\frac{1}{LC}} = \frac{1}{2\pi\sqrt{LC}}}$$

Comparison of Series and Parallel Resonant Circuit :→

Item	**Series ckt (R-L-C)**	**Parallel ckt (R– L and C)**

❖ Impedance at Resonance	Minimum	Maximum
❖ Current at Resonance	Maximum= $\frac{V}{R}$	Minimum= $\frac{V}{(L/CR)}$
❖ Effective Impedance	R	$\frac{L}{CR}$
❖ P.f. at Resonance	Unity	Unity
❖ Resonant Frequency	$\frac{1}{2\pi\sqrt{LC}}$	$\frac{1}{2\pi}\sqrt{\frac{1}{LC}-\frac{R^2}{L^2}}$
❖ It Magnifies	Voltage	Current
❖ Magnification factor	$\frac{WL}{R}$	$\frac{WL}{R}$

Parallel circuit :→



$$Z_1 = R_1 + jX_L = \sqrt{R_1^2 + X_L^2}\angle\phi_1$$

$$Z_2 = R_1 - jX_C = \sqrt{R2_1^2 + X_C^2}\angle-\phi_2$$

$$I_1 = \frac{V}{Z_1\angle\phi_1} = \frac{V}{Z_1}\angle-\phi_1 = I_1\angle-\phi_1$$

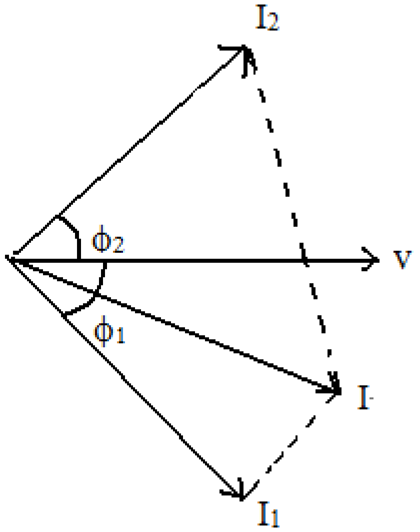
Where $\frac{V}{Z_1} = VY_1$

Here Y_1 → Admittance of the circuit

Admittance is defined as the reciprocal of impedence.

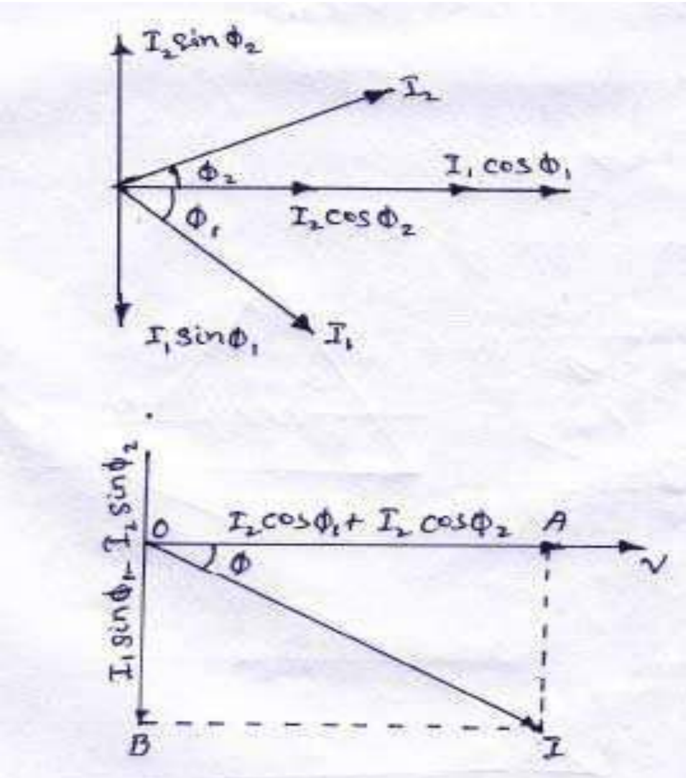
$$I_1 = VY_1 = \frac{v}{R_1 + jX_L}$$

$$I_2 = \frac{V}{Z_2\angle -\phi 2_1} = \frac{V}{Z_2}\angle\phi_1 = VY_2\angle\phi = I_2\angle\phi_2$$



$$I = \sqrt{I_1^{\,2} + I_2^{\,2} + 2I_1 I_2 \cos(\phi_1 + \phi_2)}$$

$$I = I_1\angle -\phi_1 + I_2\angle\phi_2$$



The resultant current "I" is the vector sum of the branch currents I_1 & I_2 can be found by using parallelogram low of vectors or resolving I_2 into their X

– and Y- components (or active and reactive components respectively) and then by combining these components.

Sum of active components of I_1 and I_2 = $I_1 \cos \phi_1 + I_2 \cos \phi_2$

Sum of the reactive components of I_1 and I_2 = $I_2 \sin \phi_2 - I_1 \sin \phi_1$

EXP – 01 :

A 60Hz voltage of 230 V effective value is impressed on an inductance of 0.265 H

(i) Write the time equation for the voltage and the resulting current. Let the zero axis of the voltage wave be at t = 0.

(ii) Show the voltage and current on a phasor diagram.

(iii) Find the maximum energy stored in the inductance.

Solution :-

$$V_{\max} = \sqrt{2}V = \sqrt{2} \times 230V$$

$$f = 60\text{Hz}, \quad W = 2\pi f = 2\pi \times 60 = 377 rad/s.$$

$$x_l = wl = 377 \times 0.265 = 100\Omega \quad .$$

(i) The time equation for voltage is $V(t) = 230\sqrt{2} \sin 377t.$

$$I_{\max} = V_{\max} / x_l = 230\sqrt{2}/100. = 2.3\sqrt{3}$$

$$\phi = 90^o (lag).$$

$\therefore$ Currente quation is.

$$i(t) = 2.3\sqrt{2} \sin(377t - \pi/2)$$

or $= 2.3\sqrt{2} \cos 377t$

(ii) Iti

(iii) or $E_{\max} = \frac{1}{2} LI^2_{\max} = \frac{1}{2} \times 0.265 \times (2.3\sqrt{2})^2 = 1.4J$

Example -02 :

The potential difference measured across a coil is 4.5 v, when it carries a direct current of 9 A. The same coil when carries an alternating current of 9A at 25 Hz, the potential difference is 24 v. Find the power and the power factor when it is supplied by 50 v, 50 Hz supply.

Solution :

Let R be the d.c. resistance and L be inductance of the coil.

$$R = V / I = 4.5/9 = 0.5\Omega$$

With a.c. current of 25Hz, z = V/1.

$$\frac{24}{9} = 2.66\Omega$$

$$x_l = \sqrt{Z^2 - R^2} = \sqrt{2.66^2 - 0.5^2}$$

$$= 2.62\Omega$$

$$x_l = 2\pi \times 25 \times L$$

$$x_l = 0.0167\Omega$$

At 50Hz

$$x_l = 2.62 \times 2 = 5.24\Omega$$

$$Z = \sqrt{0.5^2 + 5.24^2}$$

$$= 5.06\,\Omega$$

I = 50/5.26 = 9.5 A

P = I^2/R = $9.5^2 \times 0.5$ = 45 watt.

Example – 03 :

A 50- μf capacitor is connected across a 230-v, 50 – Hz supply. Calculate

(a) The reactance offered by the capacitor.

(b) The maximum current and

(c) The r.m.s value of the current drawn by the capacitor.

Solution :

(a) $x_l = \frac{1}{wc} = \frac{1}{2\pi \text{ fe}} = \frac{1}{2\pi \times 50 \times 50 \times 10^{-6}} = 63.6\Omega$

(c) Since 230 v represents the r.m.s value

$\because I_{rms} = 230 / x_l = 230 / 63.6 = 3.62A$

(b) $I_m = I_{r.m.s} \times \sqrt{2} = 3.62 \times \sqrt{2} = 5.11A$

Example – 04 :

In a particular R – L series circuit a voltage of 10v at 50 Hz produces a current of 700 mA. What are the values of R and L in the circuit ?

Solution :

(i) $Z = \sqrt{R^2 + (2\pi \times 50L)^2}$

$= \sqrt{R^2 + 98696L^2}$

$V = 1z$

$10 = 700 \times 10^{-3} \sqrt{(R^2 + 98696L^2)}$

$\sqrt{(R^2 + 98696L^2)} = 10 / 700 \times 10^{-3} = 100 / 7$

$R^2 + 98696L^2 = 10000/49$ ----------------------(I)

(ii) In the second case $Z = \sqrt{R^2 + (2\pi \times 75L)^2}$

$\because 10 = 500 \times 10^{-3} \sqrt{R^2 + 222066L^2)} = 20$

$\sqrt{R^2 + 222066L^2)} = 20$

CHAPTER-07
TRANSIENTS

Whenever a network containing energy storage elements such as inductor or capacitor is switched from one condition to another,either by change in applied source or change in network elements,the response current and voltage change from one state to the other state.The time taken to change from an initial steady state to the final steady state is known as the *transient period*.This response is known as *transient response* or *transients.*The response of the network after it attains a final steady value is independent of time and is called the steady-state response.The complete response of the network is determined with the help of a differential equation.

STEADY STATE AND TRANSIENT RESPONSE

In a network containing energy storage elements, with change in excitation, the currents and voltages in the circuit change from one state to other state. The behaviour of the voltage or current when it is changed from one state to another is called the transient state. The time taken for the circuit to change from one steady state to another steady state is called the transient time. The application of KVL and KCL to circuits containing energy storage elements results in differential, rather than algebraic equations. when we consider a circuit containing storage elements which are independent of the sources, the response depends upon the nature of the circuit and is called natural response. Storage elements deliver their energy to the resistances. Hence, the response changes, gets saturated after some time,and is referred to as the transient response. When we consider a source acting on a circuit, the response depends on the nature of the source or sources.This response is called forced response. In other words,the complete response of a circuit consists of two parts; the forced response and the transient response. When we consider a differential equation, the complete solution consists of two parts: the complementary function and the particular solution. The complementary function dies out after short interval, and is referred to as the transient response or source free response. The particular solution is the steady state response, or the forced response. The first step in finding the complete solution of a circuit is to form a differential equation for the circuit. By obtaining the differential equation, several methods can be used to find out the complete solution.

DC RESPONSE OF AN R-L CIRCUIT

Consider a circuit consisting of a resistance and inductance as shown in figure.The inductor in the circuit is initially uncharged and is in series with the resistor.When the switch S is closed ,we can find the complete solution for the current.Application of kirchoff's voltage law to the circuit results in the following differential equation.

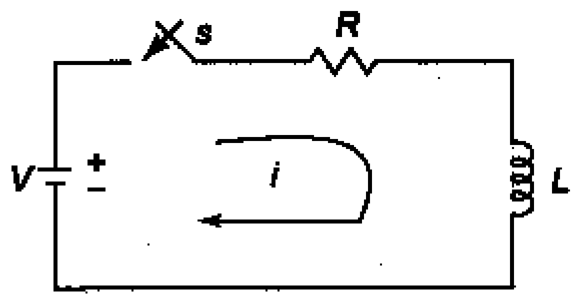


Figure 1.1

$V = Ri + L\frac{di}{dt}$ 1.1

Or $\frac{di}{dt} + \frac{R}{L}i = \frac{V}{L}$.. 1.2

In the above equation , the current I is the solution to be found and V is the applied constant voltage. The voltage V is applied to the circuit only when the switch S is closed. The above equation is a linear differential equation of first order.comparing it with a non-homogenious differential equation

$\frac{dx}{dt} + P x = K$ 1.3

whose solution is

$X = e^{-pt}\int Ke^{+pt}\,dt + c.e^{-Pt}$.. 1.4

Where c is an arbitrary constant. In a similar way , we can write the current equation as

$$i = c\, e^{-\left(\frac{R}{L}\right)t} + e^{-\left(\frac{R}{L}\right)t} \int \frac{V}{L} e^{\left(\frac{R}{L}\right)t}\,dt$$

Hence , $i = c\, e^{-\left(\frac{R}{L}\right)t} + \frac{V}{R}$.....................................1.5

To determine the value of c in equation c , we use the initial conditions .In the circuit shown in Fig.1.1, the switch s is closed at t=0.at t=0-,i.e. just before closing the switch s , the current in the inductor is zero. Since the inductor does not allow sudden changes in currents, at t=o+ just after the switch is closed,the current remains zero.

Thus at t = 0, i =0

Substituting the above condition in equation c , we have

$0 = c + \frac{V}{R}$

Substituting the value of c in equation c , we get

$$i = \frac{V}{R} - \frac{V}{R}\, e^{\frac{-Rt}{L}}$$

$$i = \frac{V}{R}\left(1 - e^{\frac{-Rt}{L}}\right)$$

$i = I_o\left(1 - e^{\frac{Rt}{L}}\right)$ (where $I_o = \frac{V}{R}$)

$i = I_o\left(1 - e^{\frac{-t}{\tau}}\right)$ (where $\tau = Time\,constant = \frac{L}{R}$) 1.6

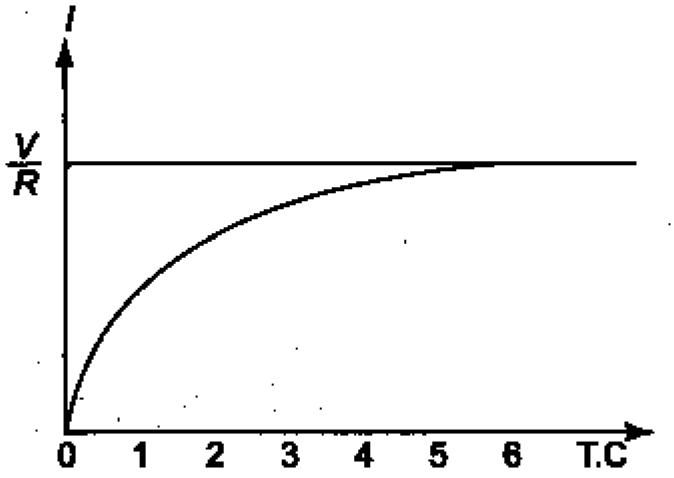


Figure 1.2

Equation d consists of two parts, the steady state part (I_o =V/R) and the transient part $I_o\, e^{\frac{-Rt}{L}}$.

When switch S is closed , the response reaches a steady state value after a time interval as shown in figure 1.2.

Here the transition period is defined as the time taken for the current to reach its final or stedy state value from its initial value.In the transient part of the solution, the quantity L/R is important in describing the curve since L/R is the time period required for the current to reach its initial value of zero to the final value I_o =V/R. The time constant of a function $I_o\, e^{\frac{-Rt}{L}}$ is the time at which the exponent of e is unity, where e is the base of the natural logarithms.The term L/R is called the time constant and is denoted by τ .

So, $\tau = \frac{L}{R}$ sec

Hence, the transient part of the solution is

$$i = -\frac{V}{R}\, e^{\frac{-Rt}{L}} = \frac{V}{R}\, e^{\frac{-t}{\tau}}$$

At one Time constant , the transient term reaches 36.8 percent of its initial value.

$$i(\tau) = -\frac{V}{R}\, e^{\frac{-t}{\tau}} = -\frac{V}{R} e^{-1} = -0.368\,\frac{V}{R}$$

Similarly,

$$i(2\tau) = -\frac{V}{R} e^{-2} = -0.135\,\frac{V}{R}$$

$$i(3\tau) = -\frac{V}{R} e^{-3} = -0.0498\,\frac{V}{R}$$

$$i(5\tau) = -\frac{V}{R} e^{-5} = -0.0067\,\frac{V}{R}$$

After 5 TC the transient part reaches more than 99 percent of its final value.

In figure A we can find out the voltages and powers across each element by using the current.

Voltage across the resistor is

$$v_R = R\,i = R \times \frac{V}{R}\left(1- e^{\frac{-Rt}{L}}\right)$$

Hence , $v_R = V\left(1- e^{\frac{-Rt}{L}}\right)$

Similarly, the voltage across the inductance is

$$v_L = L\frac{di}{dt} = L\,\frac{V}{R} \times \frac{R}{L}\, e^{\frac{-Rt}{L}} = V\, e^{\frac{-Rt}{L}}$$

The responses are shown in Figure 1.3.

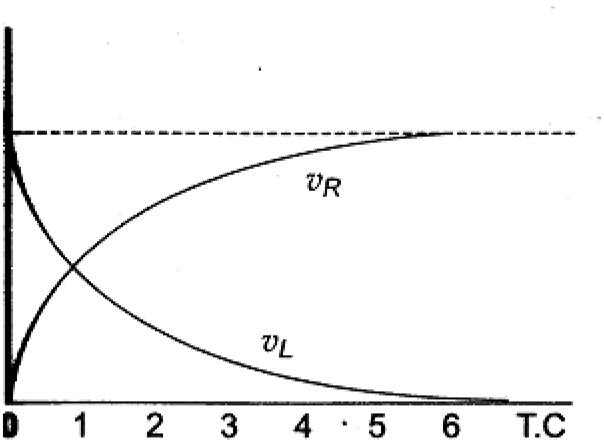


Figure 1.3

Power in the resistor is

$$P_R = v_R\, i = V\left(1- e^{\frac{-Rt}{L}}\right)\left(1 - e^{\frac{-Rt}{L}}\right) \times \frac{V}{R}$$

$$= \frac{V^2}{R}\left(1- 2e^{\frac{-Rt}{L}}\right) + e^{\frac{-2Rt}{L}}$$

Power in the inductor is

$$P_L = v_L\, i = V\, e^{\frac{-Rt}{L}} \times \frac{V}{R}\left(1 - e^{\frac{-Rt}{L}}\right)$$

$$= \frac{V^2}{R}\left(e^{\frac{-Rt}{L}} - e^{\frac{-2Rt}{L}}\right)$$

The responses are shown in figure 1.4 .

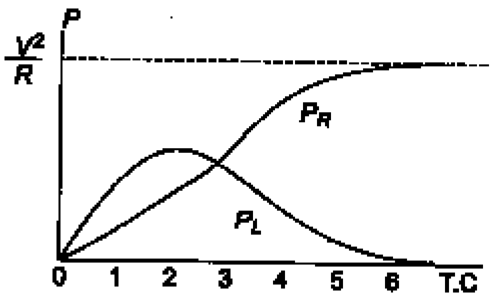


Figure 1.4

Problem : 1.1

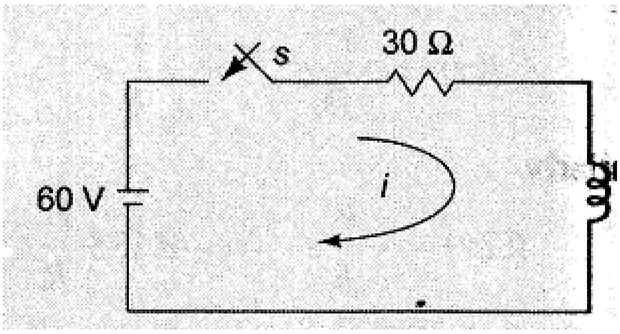


Figure 1.5

A series R-L circuit with R = 30Ω and L = 15 H has a constant voltage V = 50 V applied at t=0 as shown in Fig. 1.5 . determine the current i, the voltage across resistor and across inductor.

Solution :

By applying Kirchoff's voltage Law, we get

$15\frac{di}{dt}+30i=60$

$=> \frac{di}{dt}+2i=4$

The general solution for a linear differential equation is

$i=ce^{-pt}+e^{-pt}\int Ke^{pt}dt$

where P=2,K=4

putting the values

$i=ce^{-2t}+e^{-2t}\int 4e^{2t}dt$

$=> i=ce^{-2t}+2$

At t=0, the switch s is closed.

Since the inductor never allows sudden change in currents. At t=0^+ the current in the circuit is zero. Therefore at t=0^+, i =0

$\Rightarrow$ 0=c + 2

$\Rightarrow$ c =- 2

Substituting the value of c in the current equation, we have

i=2(1- e^{-2t}) A

voltage across resistor (V_R) =iR =2(1- e^{-2t}) x 30=60(1- e^{-2t}) v

voltage across inductor (V_L) = L$\frac{di}{dt}$ = 15 $\times$ $\frac{d}{dt}$2(1- e^{-2t}) = 30$\times$ $2e^{-2t}$ v= $60e^{-2t}$

DC RESPONSE OF AN R-C CIRCUIT

Consider a circuit consisting of a resistance and capacitance as shown in figure.The capacitor in the circuit is initially uncharged and is in series with the resistor.When the switch S is closed at t=0 , we can find the complete solution for the current.Application of kirchoff's voltage law to the circuit results in the following differential equation.

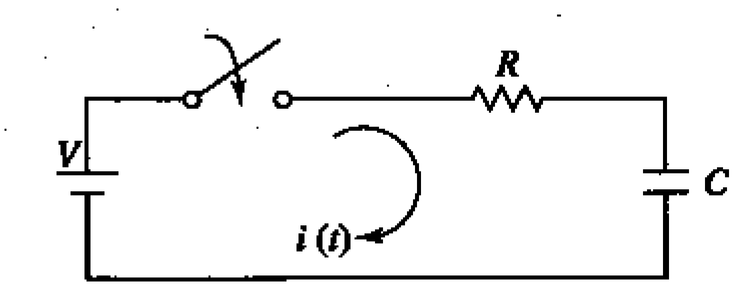


Figure 1.6

V = Ri + $\frac{1}{C}\int i\,dt$ ……………………………………………………………1.7

By differentiating the above equation, we get

0 = R $\frac{di}{dt}$ + $\frac{i}{C}$ i ……………………………………………………1.8

Or

$\frac{di}{dt}$ + $\frac{1}{RC}$ i =0 ……………………………………………………1.9

Equation c is a linear differential equation with only the complementary function. The particular solution for the above equation is zero. The solution for this type of differential equation is

$$i = c\, e^{-\left(\frac{t}{RC}\right)} \quad \text{.......................................1.10}$$

To determine the value of c in equation c , we use the initial conditions .In the circuit shown in Fig. the switch s is closed at t=0. Since the capacitor does not allow sudden changes in voltage, it will act as a short circuit at t=o+ just after the switch is closed.

So the current in the circuit at t = 0+ is $\frac{V}{R}$

Thus at t = 0, the current i = $\frac{V}{R}$

Substituting the above condition in equation c , we have

$$\frac{V}{R} = c$$

Substituting the value of c in equation c , we get

$$i = \frac{V}{R}\, e^{-\left(\frac{t}{RC}\right)} \quad \text{...1.11}$$

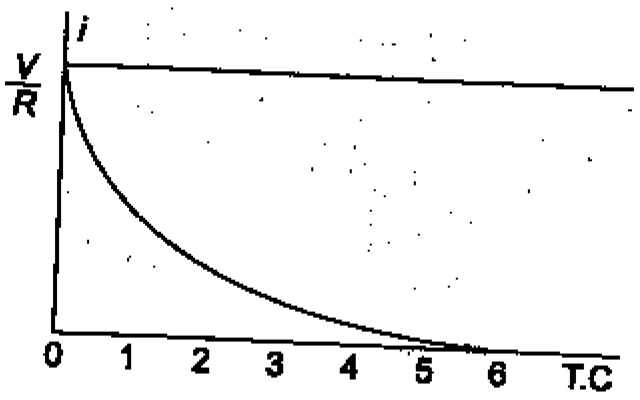


Figure 1.7

When switch S is closed , the response decays as shown in figurre.

The term RC is called the time constant and is denoted by τ .

So, $\tau = RC$ sec

After 5 TC the curve reaches 99 percent of its final value.

In figure A we can find out the voltage across each element by using the current equation.

Voltage across the resistor is

$$v_R = R\,i = R \times \frac{V}{R}\, e^{\frac{-t}{RC}}$$

Hence, $v_R = V\, e^{\frac{-t}{RC}}$

Similarly, voltage across the capacitor is

$$v_C = \frac{1}{C}\int i\, dt$$

$$= \frac{1}{C}\int \frac{V}{R} e^{\frac{-t}{RC}}\, dt$$

$$= -\left(\frac{V}{RC} \times RC\, e^{\frac{-t}{RC}}\right) + c$$

$$= -V\, e^{\frac{-t}{RC}} + c$$

At t=0,voltage across capacitor is zero

So, c = V

And

$$V_C = V\left(1 - e^{\frac{-t}{RC}}\right)$$

The responses are shown in Figure1.8.

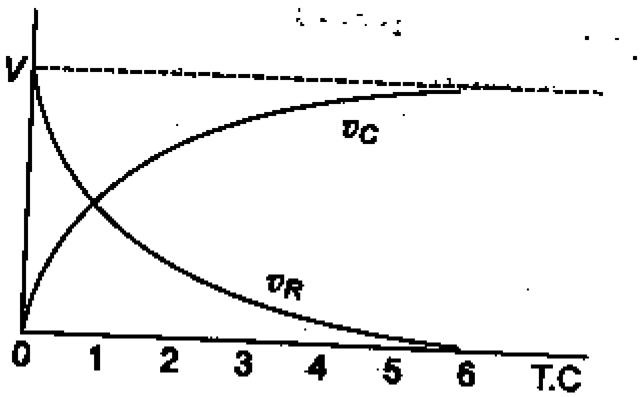


Figure 1.8

Power in the resistor is

$$P_R = v_R\, i = V\, e^{\frac{-t}{RC}} \times \frac{V}{R}\, e^{\frac{-t}{RC}}$$

$$= \frac{V^2}{R}\, e^{\frac{-2t}{RC}}$$

Power in the capacitor is

$$P_C = v_C\, i = V\left(1 - e^{\frac{-t}{RC}}\right) \frac{V}{R}\, e^{\frac{-t}{RC}}$$

$=\frac{V^2}{R}\left(e^{\frac{-t}{RC}} - e^{\frac{-2t}{RC}}\right)$

The responses are shown in figure 1.9.

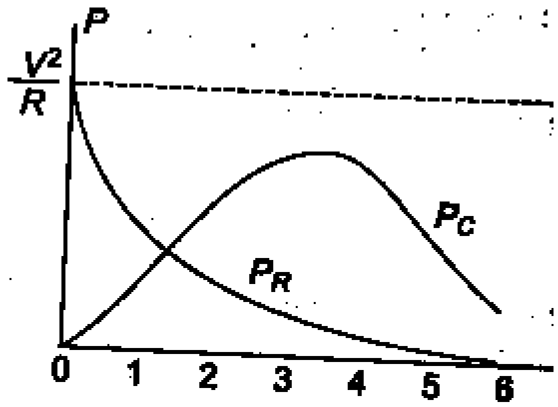


Figure 1.9

Problem : 1.2

A series R-C circuit with R = 10Ω and C =0.1 F has a constant voltage V = 20 V applied at t=0 as shown in Fig. determine the current i, the voltage across resistor and across capacitor.

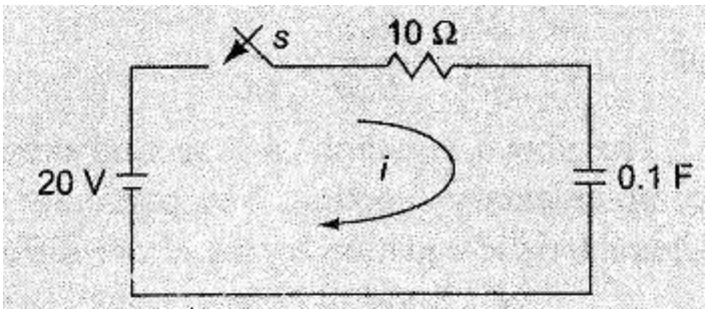


Figure 1.10

Solution :

By applying Kirchoff's voltage Law, we get

$10i + \frac{1}{0.1}\int i\, dt = 20$

Differentiating w.r.t. t we get

$10\frac{di}{dt} + \frac{i}{0.1} = 0$

$\Rightarrow \frac{di}{dt} + i = 0$

The solution for above equation is

$i = ce^{-t}$

At t=0, the switch s is closed.

Since the capacitor never allows sudden change in voltages. At t=0^+ the current in the circuit is i = V/R=20/10 =2 A

. Therefore at t= 0, i =2 A

=> the current equation is $i=2\ e^{-t}$

voltage across resistor (V_R) =iR =2 e^{-t} x 10=20 e^{-t} v

voltage across capacitor (V_C) = V$\left(1 - e^{\frac{-t}{RC}}\right)$= 20(1- e^{-t}) V

DC RESPONSE OF AN R-L-C CIRCUIT

Consider a circuit consisting of a resistance, inductance and capacitance as shown in figure.The capacitor and inductor in the circuit is initially uncharged and are in series with the resistor.When the switch S is closed at t=0 , we can find the complete solution for the current.Application of kirchoff's voltage law to the circuit results in the following differential equation.

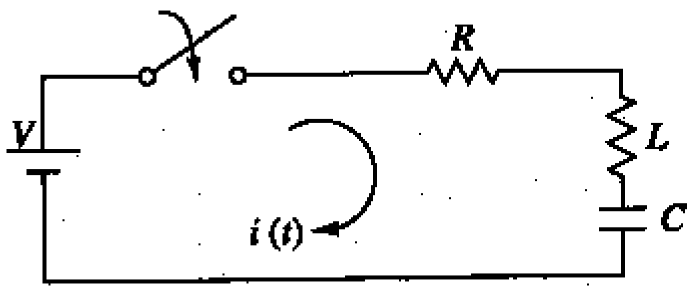


Figure 1.11

$V = Ri + L\frac{di}{dt} + \frac{1}{c}\int i\,dt$ 1.12

By differentiating the above equation, we get

$0 = R\frac{di}{dt} + L\,d^2i/dt^2 + \frac{i}{C}$ i =.. 1.13

Or

$d^2i/dt^2 + \frac{R}{L}\frac{di}{dt} + \frac{1}{LC}$ i =0.. 1.14

The above equation c is a second order linear differential equation with only the complementary function. The particular solution for the above equation is zero. The characteristics equation for this type of differential equation is

$D^2 + \frac{R}{L}D + \frac{1}{LC} = 0$ 1.15

The roots of equation 1.15 are

$$D_1, D_2 = -\frac{R}{2L} \pm \sqrt{\left(\frac{R}{2L}\right)^2 - \frac{1}{LC}}$$

By assuming $K_1 = -\frac{R}{2L}$ and $K_2 = \sqrt{\left(\frac{R}{2L}\right)^2 - \frac{1}{LC}}$

$D_1 = K_1 + K_2$ and $D_2 = K_1 - K_2$

Here K_2 may be positive,negative or zero .

Case I : K_2 *is Positive* $\left(\frac{R}{2L}\right)^2 > \frac{1}{LC}$

Then , the roots are Real and Unequal and give an over damped Response as shown in figure 1.12.

The solution for the above equation is : $i = C_1\, e^{(K_1+K_2)t} + C_2\, e^{(K_1-K_2)t}$

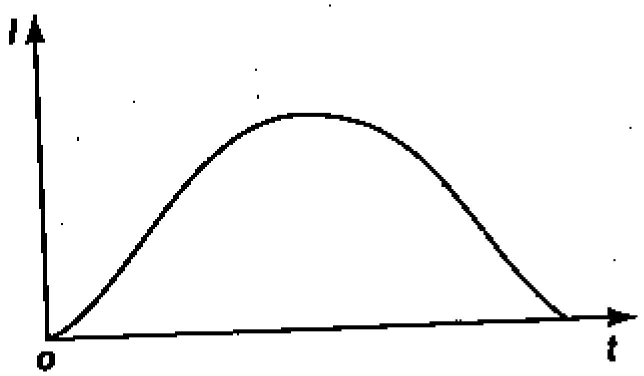


Figure 1.12

Case II : K_2 *is Negative* $\left(\frac{R}{2L}\right)^2 < \frac{1}{LC}$

Then , the roots are Complex Conjugate, and give an under-damped Response as shown in figure 1.13.

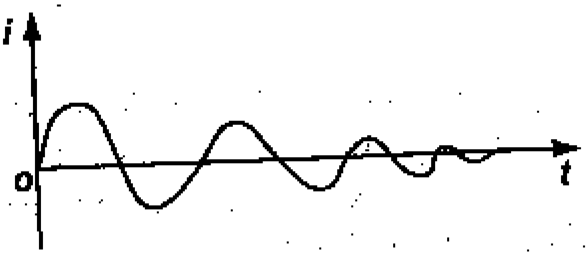


Figure 1.13

The solution for the above equation is : i = $e^{K_1 t}(C_1 \cos K_2 t + C_2 \sin K_2 t)$

Case III : K_2 *is Zero* $\left(\frac{R}{2L}\right)^2 = \frac{1}{LC}$

Then , the roots are Equal and give an Critically-damped Response as shown in figure 1.14.

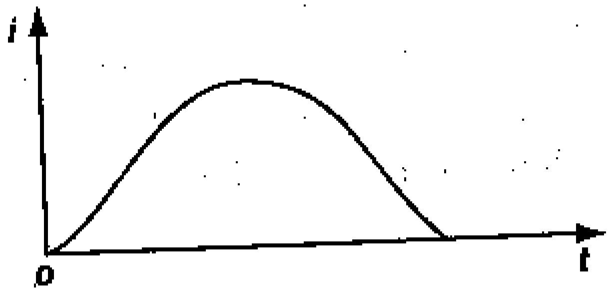


Figure 1.14

The solution for the above equation is : i = $e^{K_1 t}(C_1 + C_2 t)$

Problem : 1.3

A series R-L-C circuit with R = 20Ω , L = 0.05H and C = 20 μF has a constant voltage V = 100 V applied at t=0 as shown in Fig. determine the transient current i .

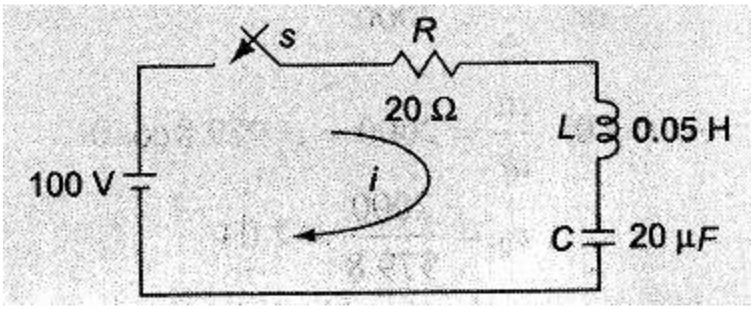


Figure 1.15

Solution :

By applying Kirchoff's voltage Law, we get

100=30i+0.05 $\frac{di}{dt} + \frac{1}{20\times10^{-6}}\int i\, dt$

Differentiating w.r.t. t we get

$0.05 d^2 i/dt^2$ +20 $\frac{di}{dt} + \frac{1}{20\times10^{-6}}$ i =0

$\Rightarrow d^2i/dt^2 + 400\frac{di}{dt} + 10^6 i = 0$

$\Rightarrow (D^2 + 400D + 10^6)i = 0$

The roots of equation are

$D_1, D_2 = -\frac{400}{2} \pm \sqrt{\left(\frac{400}{2}\right)^2 - 10^6}$

$= -200 \pm \sqrt{(200)^2 - 10^6}$

$D_1 = -200 + j979.8$

$D_2 = -200 - j979.8$

Therefore the current

$i = e^{+K_1 t}[C_1 \cos K_2 t + C_2 \cos K_2 t]$

$i = e^{-200t}[C_1 \cos 979.8t + C_2 \sin 979.8t]$ A

At t=0, the switch s is closed.

Since the inductor never allows sudden change in currents. At t= 0^+ the current in the circuit is zero. Therefore at t=0^+, i =0

$\Rightarrow$ i =0 =(1) $[C_1 \cos 0 + C_2 \sin 0]$

$\Rightarrow C_1 = 0$ and $i = e^{-200t}[C_2 \sin 979.8t]$ A

Differentiating w.r.t. t we get

$\frac{di}{dt} = C_2[e^{-200t} 979.8 \cos 979.8t + e^{-200t}(-200)\sin 979.8t]$

At t=0, the voltage across the inductor is 100 V

$\Rightarrow L\frac{di}{dt} = 100$ or $\frac{di}{dt} = 2000$

At t=0, $\frac{di}{dt} = 2000 = C_2 979.8 \cos 0$

$\Rightarrow C_2 = \frac{2000}{979.8} = 2.04$

The current equation is

i= $e^{-200t}(2.04\sin 979.8t)\, A$

ANALYSIS OF CIRCUITS USING LAPLACE TRANSFORM TECHNIQUE

The Laplace transform is a powerful Analytical Technique that is widely used to study the behavior of Linear,Lumped parameter circuits. Laplace Transform converts a time domain function f(t) to a frequency domain function F(s) and also Inverse Laplace transformation converts the frequency domain function F(s) back to a time domain function f(t).

L { f(t)} = F(s) = $\int_{-\infty}^{\infty} e^{-st}$ f(t) dt .... LT 1

L^{-1} { F(s)} = f(t) = $\frac{1}{2\pi j}\int_{-j}^{+j} F(s)\, e^{st}$ds .... LT 2

DC RESPONSE OF AN R-L CIRCUIT (LT Method)

Let us determine the solution i of the first order differential equation given by equation A which is for the DC response of a R-L Circuit under the zero initial condition i.e. current is zero, i=0 at t= 0^- and hence i=0 at t= 0^+ in the circuit in figure A by the property of Inductance not allowing the current to change as switch is closed at t=0.

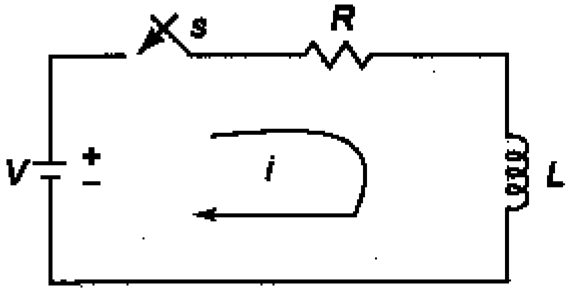


Figure LT 1.1

V = Ri + L $\frac{di}{dt}$ ....LT 1.1

Taking the Laplace Transform of bothe sides we get,

$\frac{V}{s}$ =R I(s) + L [s I(s) –I(0)] ....LT 1.2

=> $\frac{V}{s}$ =R I(s) + L [s I(s)] (I(0) =0 : zero initial current)

=> $\frac{V}{s}$ = I(s)[R +L s]

=>I(s) = $\frac{V}{s[R+Ls]}$ ....LT 1.3

Taking the Laplace Inverse Transform of both sides we get,

=> $\mathcal{L}^{-1}\{I(s)\} = i(t) = \mathcal{L}^{-1}\{\frac{V}{s[R+Ls]}\}$

i(t) = $\mathcal{L}^{-1}\{\frac{V/L}{s[R/L+s]}\}$ (Dividing the numerator and denominator by L)

putting $\propto = R/L$ we get

i(t) = $\mathcal{L}^{-1}\{\frac{V/L}{s[s+\alpha]}\}$ = $\mathcal{L}^{-1}\{\frac{V}{L}(\frac{1}{s}-\frac{1}{(s+\alpha)})\frac{1}{\alpha}\}$

i(t) = $\mathcal{L}^{-1}\{\frac{V}{L}(\frac{1}{s}-\frac{1}{(s+R/L)})\frac{L}{R}\}$ (again putting back the value of $\propto$)

i(t) = $\mathcal{L}^{-1}\{\frac{V}{R}(\frac{1}{s}-\frac{1}{(s+R/L)})\}$ = $\frac{V}{R}$(1- $e^{\frac{-Rt}{L}}$) = I_o (1- $e^{\frac{-Rt}{L}}$) (where $I_o = \frac{V}{R}$)

i(t)= I_o (1- $e^{\frac{-t}{\tau}}$) (where $\tau = Time\ constant = \frac{L}{R}$) .... LT 1.4

It can be observed that solution for i(t) as obtained by Laplace Transform technique is same as that obtained by standard differential method .

DC RESPONSE OF AN R-C CIRCUIT(L.T.Method)

Similarly ,

Let us determine the solution i of the first order differential equation given by equation A which is for the DC response of a R-C Circuit under the zero initial condition i.e. voltage across capacitor is zero, V_c =0 at t=0^- and hence V_c =0 at t=0^+ in the circuit in figure A by the property of capacitance not allowing the voltage across it to change as switch is closed at t=0.

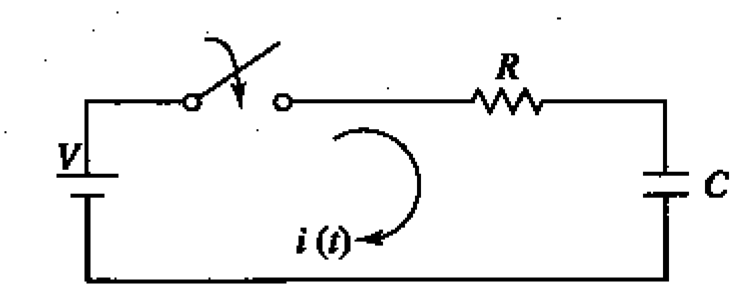


Figure LT 1.2

V = Ri + $\frac{1}{C}\int i\,dt$ ....LT 1.5

Taking the Laplace Transform of both sides we get,

$\frac{V}{s}$ =R I(s) + $\frac{1}{C}$[$\frac{I(s)}{s}$ +I (0)] .... LT 1.6

=> $\frac{V}{s}$ =R I(s) + $\frac{1}{C}$[$\frac{I(s)}{s}$] (I(0) =0 : zero initial charge)

=> $\frac{V}{s}$ = I(s)[R +$\frac{1}{Cs}$] = I(s)[$\frac{RCs+1}{Cs}$]

=>I(s) = $\frac{V}{s}\left[\frac{Cs}{(RCs+1)}\right] = \frac{VC}{(RCs+1)}$ ....LT 1.7

Taking the Laplace Inverse Transform of both sides we get,

=> $L^{-1}\{I(s)\} = i(t) = L^{-1}\{\frac{VC}{(RCs+1)}\}$

i(t) = $L^{-1}\{\frac{\frac{VC}{RC}}{[s+\frac{1}{RC}]}\}$ (Dividing the numerator and denominator by RC)

putting $\propto = \frac{1}{RC}$ we get

i(t) = $L^{-1}\{\frac{V/R}{[s+\propto]}\} = \frac{V}{R}\, e^{-\propto t}$

i(t) = $\frac{V}{R}\, e^{\frac{-t}{RC}}$ (putting back the value of $\propto$)

i(t) = $I_o e^{\frac{-t}{RC}}$ (where $I_o = \frac{V}{R}$) ....LT 1.8

i(t)= $I_o\, e^{\frac{-t}{\tau}}$) (where $\tau = Time\ constant =$ RC)

It can be observed that solution for i(t) as obtained by Laplace Transform technique in q is same as that obtained by standard differential method in d.

DC RESPONSE OF AN R-L-C CIRCUIT (L.T. Method)

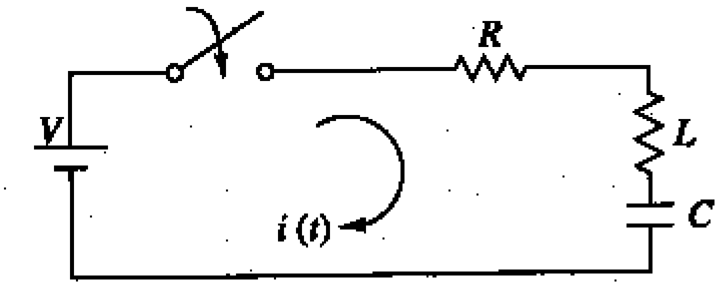


Figure LT 1.3

Similarly ,

Let us determine the solution i of the first order differential equation given by equation A which is for the DC response of a R-L-C Circuit under the zero initial condition i.e. the switch s is closed at t=0.at t=0-,i.e. just before closing the switch s , the current in the inductor is zero. Since the inductor does not allow sudden changes in currents, at t=o+ just after the switch is closed,the current remains zero. also the voltage across capacitor is zero i.e. V_c =0 at t= 0^- and hence V_c =0 at t= 0^+ in the circuit in figure by the property of capacitance not allowing the voltage across it V_c to suddenly change as switch is closed at t=0.

V = Ri + L $\frac{di}{dt} + \frac{1}{C}\int i\,dt$...................................... LT 1.9

Taking the Laplace Transform of both sides we get,

$\frac{V}{s}$ =R I(s) ++ L [s I(s) –I(0)]+$\frac{1}{C}$[$\frac{I(s)}{s}$ +I (0)]...................................... LT 1.10

=> $\frac{V}{s}$ =R I(s) + L [s I(s)] + $\frac{1}{C}$[$\frac{I(s)}{s}$]   (I(0) = 0: zero initial current & I(0) =0 : zero initial charge)

=> $\frac{V}{s}$ = I(s)[R +Ls + $\frac{1}{Cs}$] = I(s)[$\frac{LCs^2+Rcs+1}{Cs}$]

=>I(s) = $\frac{V}{s}$[$\frac{Cs}{(LCS^2+RCs+1)}$] =$\frac{VC}{(LCs^2+RCs+1)}$ ....LT 1.11

Taking the Laplace Inverse Transform of both sides we get,

=> L^{-1}{I(s)} = $i(t)$ = $L^{-1}\{\frac{VC}{(LCs^2+RCs+1)}\}$

i(t) = $L^{-1}\{\frac{\frac{VC}{LC}}{[s^2+s\frac{RC}{LC}+\frac{1}{LC}]}\}$ (Dividing the numerator and denominator by LC)

i(t) = $L^{-1}\{\frac{\frac{V}{L}}{[s^2+s\frac{R}{L}+\frac{1}{LC}]}\}$

putting $\propto = \frac{R}{2L}$ and $\omega = \sqrt{\frac{1}{LC}}$ we get

i(t) = $L^{-1}\{\frac{\frac{V}{L}}{[s^2+2\propto s+\omega^2]}\}$

The denominator polynomial becomes = $[s^2 + 2 \propto s + \omega^2]$

where, s_1 , s_2 = $\frac{-2\propto\pm\sqrt{4\propto^2-4\omega^2}}{2}$ = $-\propto\pm\sqrt{\propto^2-\omega^2}$ = $-\propto\pm\beta$

where, $\propto = \frac{R}{2L}$; $\omega = \sqrt{\frac{1}{LC}}$ and $\beta = \sqrt{\propto^2-\omega^2}$

By partial Fraction expansion , of I(s) ,

$$I(s) = \frac{A}{s-s_1} + \frac{B}{s-s_2}$$

$$A = (s - s_1)\ I(s)]\ s=s_1$$

$$= \frac{\frac{V}{L}}{(s_1-s_2)}$$

$$B = (s - s_2)\ I(s)]\ s=s_2$$

$$= \frac{\frac{V}{L}}{(s_2-s_1)} = -\frac{\frac{V}{L}}{(s_1-s_2)}$$

$$I(s) = \frac{\frac{V}{L}}{(s_1-s_2)}\left(\frac{1}{(s-s_1)} - \frac{1}{(s-s_2)}\right)$$

Taking the Inverse Laplace Transform

$$i(t) = A_1 e^{s_1 t} + A_2 e^{s_2 t}$$

Where A_1 and A_2 are constants to be determined and s_1 and s_2 aren the roots of the equation.

Now depending upon the values of s_1 and s_2 , we have three cases of the response.

CASE I : When the roots are Real and Unequal, it gives an over-damped response.

$\frac{R}{2L} > \sqrt{\frac{1}{LC}}$ or $\alpha > \omega$; In this case, the solution is given by

$$i(t) = e^{-\alpha t} (A_1 e^{\beta t} + A_2 e^{-\beta t}) \text{ .... LT 1.12}$$

or $$i(t) = A_1 e^{s_1 t} + A_2 e^{s_2 t} \quad \text{for } t > 0$$

CASE II : When the roots are Real and Equal, it gives an Critically-damped response.

$\frac{R}{2L} = \sqrt{\frac{1}{LC}}$ or $\alpha = \omega$; In this case, the solution is given by

or

$$i(t) = e^{-\alpha t} (A_1 + A_2 t) \quad \text{for } t > 0 \text{ ....LT 1.13}$$

CASE III : When the roots are Complex Conjugate, it gives an under-damped response.

$\frac{R}{2L} < \sqrt{\frac{1}{LC}}$ or $\alpha < \omega$; In this case, the solution is given by

$$i(t) = A_1 e^{s_1 t} + A_2 e^{s_2 t} \quad \text{for } t > 0$$

where, $s_1, s_2 = \frac{-2\alpha \pm \sqrt{4\alpha^2 - 4\omega^2}}{2} = -\alpha \pm \sqrt{\alpha^2 - \omega^2}$

Let $\sqrt{\alpha^2 - \omega^2} = \sqrt{-1}\sqrt{\omega^2 - \alpha^2} = j\,\omega_d$ where $j = \sqrt{-1}$ and $\omega_d = \sqrt{\omega^2 - \alpha^2}$

Hence , $$i(t) = e^{-\alpha t}(A_1 e^{j\omega_d t} + A_2 e^{-j\omega_d t})$$

$$i(t) = e^{-\alpha t} \left[(A_1 + A_2) \left\{ \frac{e^{j\omega_d t} + e^{-j\omega_d t}}{2} \right\} + j(A_1 - A_2) \left\{ \frac{e^{j\omega_d t} - e^{-j\omega_d t}}{2j} \right\} \right]$$

$$i(t) = e^{-\alpha t} \left[(A_1 + A_2)\cos\omega_d t + j(A_1 - A_2)\sin\omega_d t \right]$$

$$i(t) = e^{-\alpha t}(B_1 \cos\omega_d t + B_2 \sin\omega_d t) \text{ ....LT 1.14}$$

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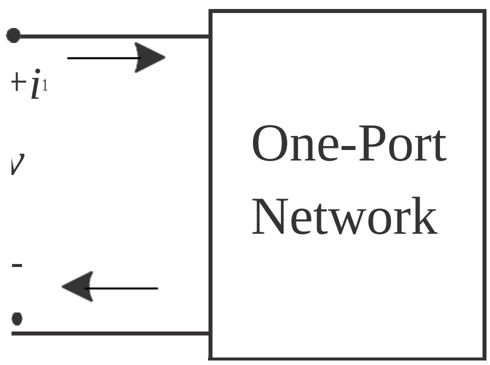
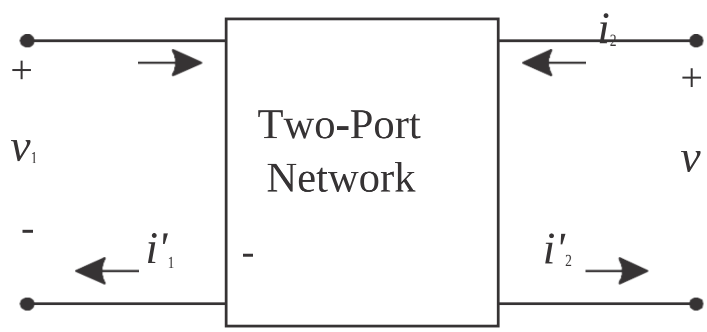
CHAPTER 8

TWO-PORT NETWORKS



a) **One port network** is a two terminal electrical network in which, current enters through one terminal and leaves through another terminal. Resistors, inductors and capacitors are the examples of one port network because each one has two terminals. One port network representation is shown in the following figure.
b) A pair of terminals at which a signal (voltage or current) may enter or leave is called a port.
c) A network having only one such pair of terminals is called a one-port network.
d) No connections may be made to any other nodes internal to the network.
e) By KCL, we therefore havei_1=i_1



- **two port network** is a pair of two terminal electrical network in which, current enters through one terminal and leaves through another terminal of each port. Two port network representation is shown in the following figure.Type equation here.
- Two-port networks are used to describe the relationship between a pair of terminals
- The analysis methods we will discuss require the following conditions be met
    1. Linearity
    2. No independent sources inside the network
    3. No stored energy inside the network (zero initial conditions)
    4. $i_1 = i_1$ and $i_2 = i$

Two Port Network Parameters

There are various parameters needed to analyze a two port network. For examples, Z parameters, Y parameters, h parameters, g parameters, ABCD parameters etc.
Let us discuss these network parameters one by one to gain a better understanding of their application and uses.

Impedance Parameters

- Suppose the currents and voltages can be measured.
- Alternatively, if the circuit in the box is known, V_1 and V_2 can be calculated based on circuit analysis.
- Relationship can be written in terms of the impedance parameters.
- We can also calculate the impedance parameters after making two sets of measurements.

$$V_1 = z_{11} I_1 + z_{12} I_2$$

$$V_2 = z_{21} I_1 + z_{22} I_2$$

If the right port is an open circuit ($I_2=0$), then we can easily solve for two of the impedance parameters: Similarly by open circuiting left hand port ($I_{1=0}$) we can solve for the other two parameters.

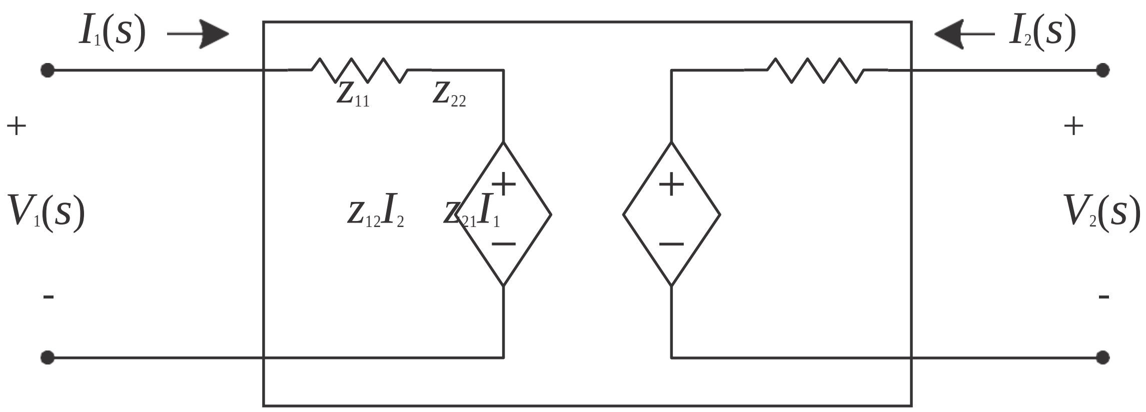
2

$$Z_{11} = inputimpedence = \frac{V_1}{I_1} \Big| I_2 = 0 \qquad Z_{21} = forwardtransferimpedence = \frac{V_2}{I_1} \Big| I_2 = 0$$

$$Z_{12} = reversetransferimpedence = \frac{V_1}{I_2} \Big| I_1 = 0 \qquad Z_{22} = outputimpedence = \frac{V_2}{I_2} \Big| I_1 = 0$$

—| —|

Impedance Parameter Equivalent

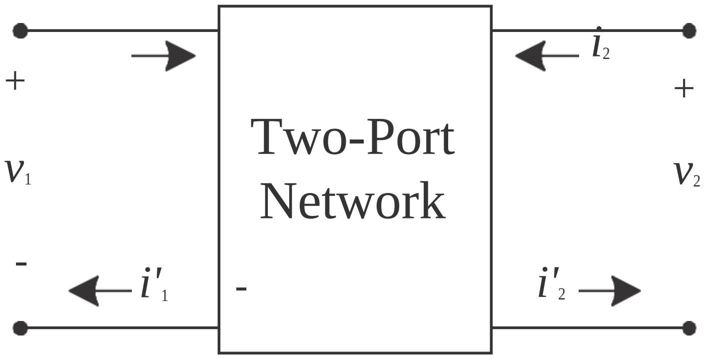


$$V_1 = z_{11}I_1 + z_{12}I_2$$

$$V_2 = z_{21}I_1 + z_{22}I_2$$

- Once we know what the impedance parameters are, we can model the behavior of the two-port with an equivalent circuit.
- Notice the similarity to Th´evenin and Norton equivalents

Admittance Parameters



$I_1 = y_{11}V_1 + y_{12}V_2$

$I_2 = y_{21}V_1 + y_{22}V_2$

Y_{11} = input admittance = $\frac{I_1}{V_1}\Big|V_2 = 0$

Y_{21} = forward transfer admittance = $\frac{I_2}{V_1}\Big|V_2 = 0$

Y_{22} = output admittance = $\frac{I_1}{V_2}\Big|V_1 = 0$

Y_{12} = reverse transfer admittance = $\frac{I_1}{V_2}\Big|V_1 = 0$

Hybrid Parameters

$V_1 = h_{11}I_1 + h_{12}V_2$

$I_2 = h_{21}I_1 + h_{22}V_2$

h_{11} = input impedance = $\frac{V_1}{I_1}\Big|V_2 = 0$

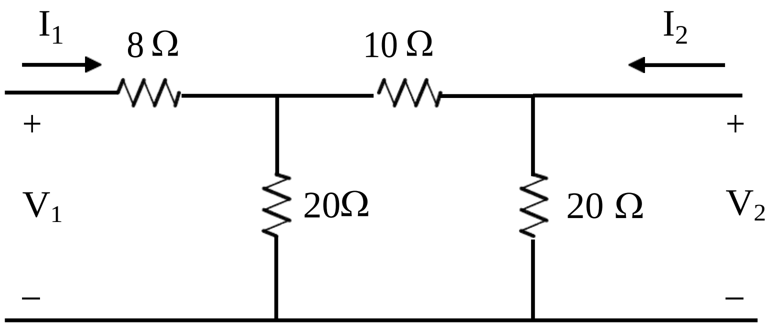
h_{21} = forward current ratio = $\frac{I_2}{I_1}\Big|V_2 = 0$

h_{12} = reverse voltage ratio = $\frac{V_1}{V_2}\Big|I_1 = 0$

h_{22} = output admittance = $\frac{I_2}{V_2}\Big|I_1 = 0$

Example:

Given the following circuit. Determine the Z parameters.



$Z_{11} = 8 + 20 \| 30 = 20\ \Omega$

$Z_{22} = 20 \| 30 = 12\ \Omega$

$$Z_{12} = \frac{V_1}{I_2}\Big|_{I_1 = 0}$$

$$V_1 = \frac{20xI_2x20}{20+30} = 8xI_2 \quad \text{Therefore } z_{12} = \frac{8xI_2}{I_2} = 8\ \Omega = z_{21}$$

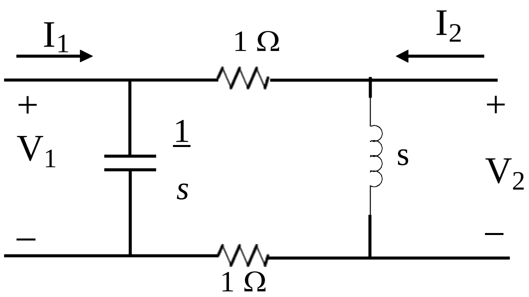
The Z parameter equations can be expressed in matrix form as follows.

$$\begin{bmatrix} V_1 \\ V_2 \end{bmatrix} = \begin{bmatrix} z_{11} & z_{12} \\ z_{21} & z_{22} \end{bmatrix} \begin{bmatrix} I_1 \\ I_2 \end{bmatrix}$$

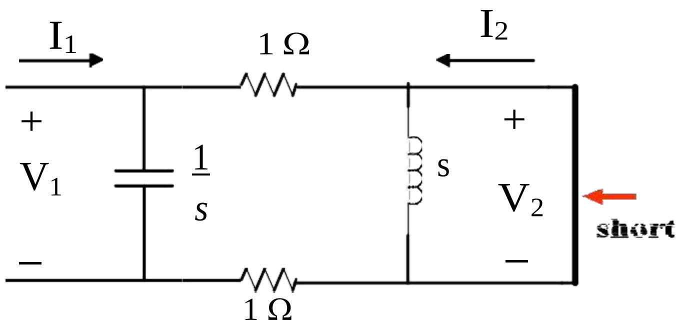
$$\begin{bmatrix} V_1 \\ V_2 \end{bmatrix} = \begin{bmatrix} 20 & 8 \\ 8 & 12 \end{bmatrix} \begin{bmatrix} I_1 \\ I_2 \end{bmatrix}$$

Example:

Given the following circuit. Determine the Y parameters.



$$I_1 = y_{11}V_1 + y_{12}V_2$$
$$I_2 = y_{21}V_1 + y_{22}V_2$$



To find y_{11}

$$V_1 = I_1\left(\frac{2/s}{2+1/s}\right) = I_1\left[\frac{2}{2s+1}\right]$$

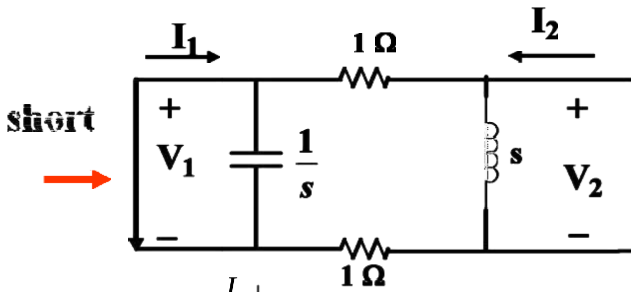
So $y_{11} = \frac{I}{V_1}$ $$y_{11} = \frac{I_1}{V_1}\Big|_{V_2=0} = s + 0.5$$

To find y_{12} and y_{21} we reverse things and short V_1

$$y_{21} = \frac{I_2}{V_1}\Big|_{V_2=0}$$

$$V_1 = -2I_2$$

$$y_{21} = \frac{I_2}{V_1} = 0.5 \text{ S}$$



$$y_{12} = \frac{I_1}{V_2}\Big|_{V_1=0}$$

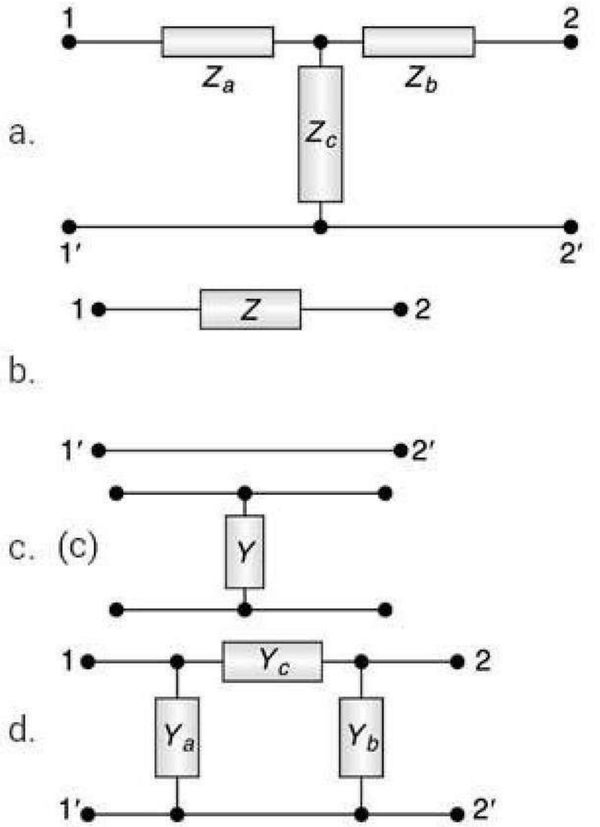
$$V_2 = -2I_1 \qquad y_{12} = \frac{I_1}{V_2} = 0.5s$$

$$y_{22} = 0.5 + \frac{1}{s}$$

$$y_{22} = \frac{I_2}{V_2}\Big|_{V_1=0} \qquad V_2 = I_2\frac{2s}{(s+2)} \qquad y_{22} = 0.5 + \frac{1}{s}$$

Problem 1

Find the Z and Y parameter for the networks shown in figure.



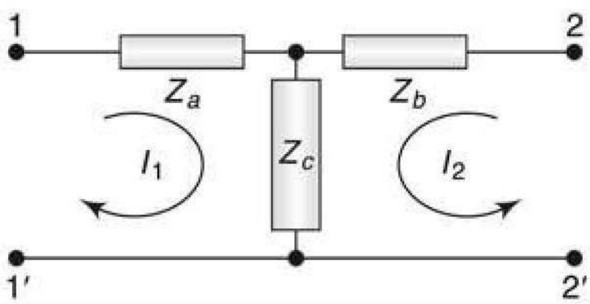
Solution

a. By KVL, $(Z_a + Z_c)I_1 + Z_c I_2 = V_1$

and $Z_c I_1 + (Z_b + Z_c)I_2 = V_2$

Thus, the Z-parameters are:

$z_{11} = (Z_a + Z_c)$, $z_{12} = z_{21} = Z_c$, $z_{22} = (Z_b + Z_c)$



b. By KCL,

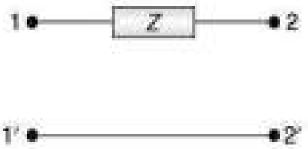
$$I_1 = \frac{V_1 - V_2}{Z} = \frac{1}{Z}V_1 - \frac{1}{Z}V_2$$

and $$I_2 = \frac{V_2 - V_1}{Z} = -\frac{1}{Z}V_1 + \frac{1}{Z}V_2$$

Thus, the y-parameters are,

$$y_{11} = \frac{1}{Z} = y_{22} \quad y_{12} = y_{21} = -\frac{1}{Z}$$

Since, $\Delta y = y_{11}y_{22} - y_{12}y_{21} = 0$, the z-parameters do not exist for this network.



c. By KVL,

$V_1 =$

$$\frac{I_1 + I_2}{Y} = V_2 \quad \text{or, } V_1 = \left(\frac{1}{Y}\right)I_1 + \left(\frac{1}{Y}\right)I_2 \quad \text{and} \quad V_2 = \left(\frac{1}{Y}\right)I_1 + \left(\frac{1}{Y}\right)I_2$$

Thus, the z-parameters are,

$$z_{11} = z_{22} = \frac{1}{Y} = z_{12} = z_{21}$$

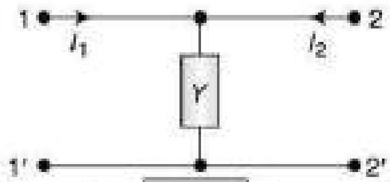
Since, $\Delta z = z_{11}z_{22} - z_{12}z_{21} = 0$, the y-parameters do not exist for this network.

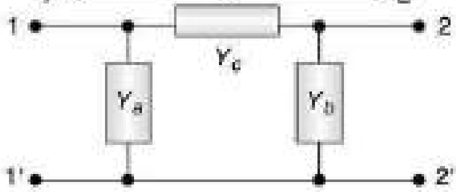
d. By KCL,

$$I_1 = Y_aV_1 + (V_1 - V_2)Y_c = V_1(Y_a + Y_c) - V_2Y_c$$
$$I_2 = Y_bV_2 + (V_2 - V_1)Y_c = -V_1Y_c + V_2(Y_b + Y_c)$$

Thus, the y-parameters are:

$$y_{11} = Y_a + Y_c;\ y_{12} = y_{21} = -Y_c;\ y_{22} = Y_b + Y_c$$





Problem 2

a. The following equations give the voltages V_1 and V_2 at the two ports of a two port network, $V_1 = 5I_1 + 2I_2$, $V_2 = 2I_1 + I_2$;

   A load resistance of 3 Ω is connected across port-2. Calculate the input impedance.

b. The *z*-parameters of a two port network are $z_{11} = 5\ \Omega$, $z_{22} = 2\ \Omega$, $z_{12} = z_{21} = 3\ \Omega$. Load resistance of 4 Ω is connected across the output port. Calculate the input impedance.

Solution

a. From the given equations,

$V_1 = 5I_1 + 2I_2$ (i)

$V_2 = 2I_1 + I_2$ (ii)

At the output, $V_2 = -I_2R_L = -3I_2$

Putting this value in (ii),

$-3I_2 = 2I_1 + I_2$ fi $I_2 = -I_1/2$

Putting in (i), $V_1 = 5I_1 + \left(\frac{-I_1}{2}\right) = 4I_1$

$\therefore$ Input impedance, $Z_{in} = \frac{V_1}{I_1} = 4\Omega$

b. [Same as Prob. (a)] $Z_{in} = \frac{V_1}{I_1} = 3.5\Omega$

Problem 3

Determine the h-parameter with the following data:

i. with the output terminals short circuited, V_1 = 25 V, I_1 = 1 A, I_2 = 2 A

ii. with the input terminals open circuited, V_1 = 10 V, V_2 = 50 V, I_2 = 2 A

Solution

The h-parameter equations are,

$$V_1 = h_{11}I_1 + h_{12}V_2$$

$$I_2 = h_{21}I_1 + h_{22}V_2$$

a. With output short-circuited, $V_2 = 0$, given: V_1 = 25 V, I_1 = 1 A and I_2 = 2 A.

$$\begin{aligned} \therefore \quad & 25 = h_{11} \times 1 \\ \text{and} \quad & 2 = h_{21} \times 1 \end{aligned} \Bigg\} \Rightarrow h_{11} = 25\ \Omega, \text{ and } h_{21} = 2$$

b. With input open-circuited, $I_1 = 0$, given: V_1 = 10 V, V_2 = 50 V and I_2 = 2 A.

$$\begin{aligned} \therefore \quad & 10 = h_{12} \times 50 \\ \text{and} \quad & 2 = h_{22} \times 50 \end{aligned} \Bigg\} \Rightarrow h_{12} = \frac{1}{5} = 0.2 \text{ and } h_{23} = \frac{1}{25}\ ℧ = 0.04\ ℧$$

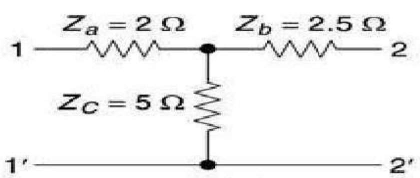
Thus, the h-parameters are:

$$[h] = \begin{bmatrix} 25\ \Omega & 0.2 \\ 2 & 0.04\ \Omega^{-1} \end{bmatrix}$$

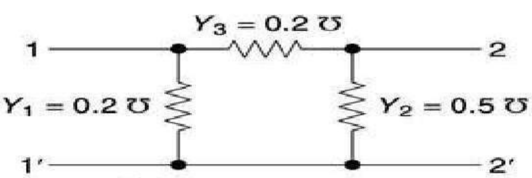
Problem 4

a. Find the equivalent π-network for the T-network shown in the Fig. (a).

b. Find the equivalent T-network for the π-network shown in the Fig. (b).



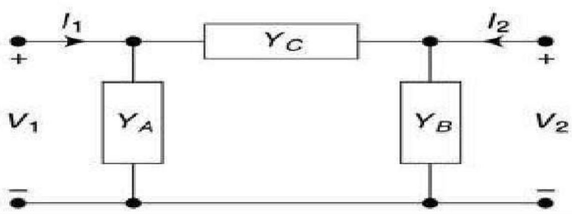
(a)



(b)

Solution

a. Let the equivalent π-network have Y_C as the series admittance and Y_A and Y_B as the shunt admittances at port-1 and port-2, respectively.



Now, the z-parameters are given as:

$z_{11} = (Z_A + Z_C) = 7\ \Omega$, $z_{12} = z_{21} = Z_C = 5\ \Omega$, $z_{22} = (Z_B + Z_C) = 7.5\ \Omega$

$\therefore \Delta z = (7 \times 7.5 - 5 \times 5) = 27.5\ \Omega^2$

$$\therefore y_{11} = \frac{z_{22}}{\Delta z} = \frac{7.5}{27.5}\ \mho$$

$$y_{12} = y_{21} = -\frac{z_C}{\Delta z} = -\frac{5}{27.5}\ \mho$$

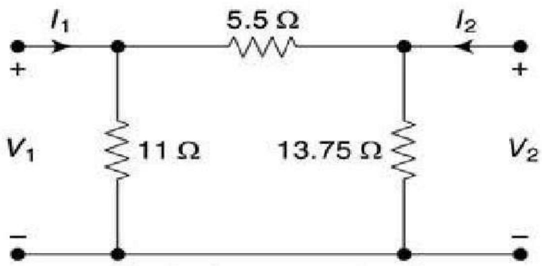
$$y_{22} = \frac{z_{11}}{\Delta z} = \frac{7}{27.5}\ \mho$$

$$\therefore Y_A = (y_{11} + y_{12}) = \frac{2.5}{27.5} = \frac{1}{11}\ \mho$$

$$\therefore Y_B = (y_{22} + y_{12}) = \frac{2}{27.5}\ \mho$$

$$\text{and } Y_C = -y_{21} = \frac{5}{27.5} = \frac{2}{11}\ \mho$$

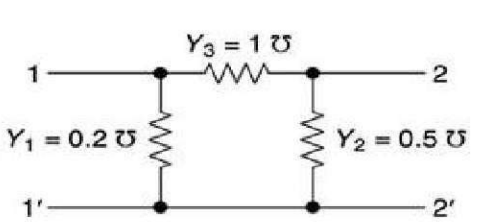
Thus, the impedances of the equivalent π-networks are:



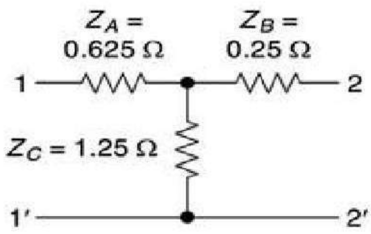
Equivalent π-network

$$\left.\begin{aligned} Z_A &= \frac{1}{Y_A} = 11\ \Omega, \\ Z_B &= \frac{1}{Y_B} = 13.75\ \Omega, \\ Z_C &= \frac{1}{Y_C} = 5.5\ \Omega \end{aligned}\right\}$$

b.



π-network



Equivalent T-network

The y-parameters,

$$y_{11} = 1.2\ \mho,\ y_{12} = y_{21} = -1\ \mho, \text{ and } y_{22} = 1.5\ \mho$$

$\therefore \Delta y = (1.2 \times 1.5 - 1) = 0.8$

$$\therefore z_{11} = \frac{y_{22}}{\Delta y} = \frac{1.5}{0.8}\ \Omega,\ z_{12} = z_{21} = -\frac{y_{12}}{\Delta y} = \frac{1}{0.8}\ \Omega,\ z_{22} = \frac{y_{11}}{\Delta y} = \frac{1.2}{0.8}\ \Omega$$

$$\left.\begin{aligned} Z_A &= (z_{11} - z_{12}) = \frac{0.5}{0.8} = 0.625\ \Omega \\ \therefore Z_B &= (z_{22} - z_{12}) = \frac{0.2}{0.8} = 0.25\ \Omega \\ Z_C &= z_{12} = \frac{1}{0.8} = 1.25\ \Omega \end{aligned}\right\}$$

CHAPTER 9

LOW PASS FILTER INTRODUCTION

Basically, an electrical filter is a circuit that can be designed to modify, reshape or reject all unwanted frequencies of an electrical signal and accept or pass only those signals wanted by the circuit's designer. In other words they "filter-out" unwanted signals and an ideal filter will separate and pass sinusoidal input signals based upon their frequency.

In low frequency applications (up to 100kHz), passive filters are generally constructed using simple RC(Resistor-Capacitor) networks, while higher frequency filters (above 100kHz) are usually made from RLC (Resistor-Inductor-Capacitor) components.

Passive Filters are made up of passive components such as resistors, capacitors and inductors and have no amplifying elements (transistors, op-amps, etc) so have no signal gain, therefore their output level is always less than the input.

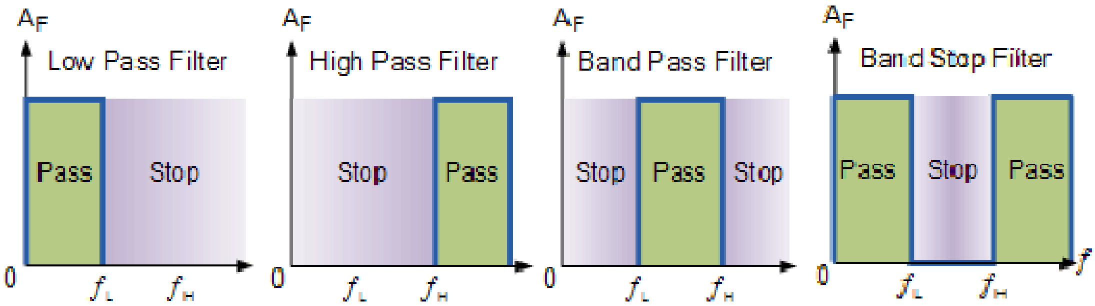
Filters are so named according to the frequency range of signals that they allow to pass through them, while blocking or "attenuating" the rest. The most commonly used filter designs are the:

- 1. The Low Pass Filter – the low pass filter only allows low frequency signals from 0Hz to its cut-off frequency, ƒc point to pass while blocking those any higher.
- 2. The High Pass Filter – the high pass filter only allows high frequency signals from its cut-off frequency, ƒc point and higher to infinity to pass through while blocking those any lower.
- 3. The Band Pass Filter – the band pass filter allows signals falling within a certain frequency band setup between two points to pass through while blocking both the lower and higher frequencies either side of this frequency band.
- 4 Band Stop Filter - It is so called *band-elimination*, *band-reject*, or *notch* filters; this kind of filter passes all frequencies above and below a particular range set by the component values.

Simple First-order passive filters (1st order) can be made by connecting together a single resistor and a single capacitor in series across an input signal, (Vin) with the output of the filter, (Vout) taken from the junction of these two components. Depending on which way around we connect the resistor and the capacitor with regards to the output signal determines the type of filter construction resulting in either a Low Pass Filter or a High Pass Filter.

As the function of any filter is to allow signals of a given band of frequencies to pass unaltered while attenuating or weakening all others those are not wanted, we can define the amplitude response characteristics of an ideal filter by using an ideal frequency response curve of the four basic filter types as shown.

IDEAL FILTER RESPONSE CURVES

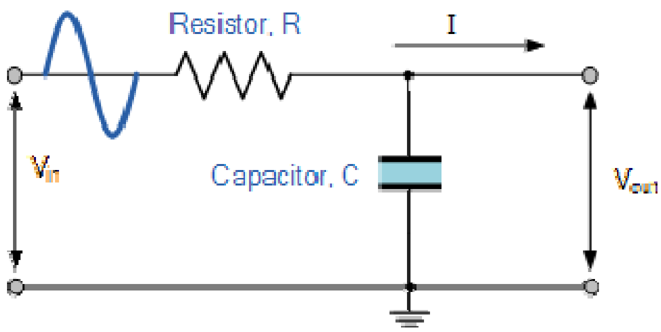


A Low Pass Filter can be a combination of capacitance, inductance or resistance intended to produce high attenuation above a specified frequency and little or no attenuation below that frequency. The frequency at which the transition occurs is called the "cutoff" frequency. The simplest low pass filters consist of a resistor and capacitor but more sophisticated low pass filters have a combination of series inductors and parallel capacitors. In this tutorial we will look at the simplest type, a passive two component RC low pass filter.

THE LOW PASS FILTER

A simple passive RC Low Pass Filter or LPF, can be easily made by connecting together in series a single Resistor with a single Capacitor as shown below. In this type of filter arrangement the input signal (Vin) is applied to the series combination (both the Resistor and Capacitor together) but the output signal (Vout) is taken across the capacitor only. This type of filter is known generally as a "first-order filter" or "one-pole filter", why first-order or single-pole?, because it has only "one" reactive component, the capacitor, in the circuit.

RC LOW PASS FILTER CIRCUIT



As mentioned previously in the Capacitive Reactance tutorial, the reactance of a capacitor varies inversely with frequency, while the value of the resistor remains constant as the frequency changes. At low frequencies the capacitive reactance, (Xc) of the capacitor will be very large compared to the resistive value of the resistor, R and as a result the voltage across the capacitor, Vc will also be large while the voltage drop across the resistor, Vr will be much lower. At high frequencies the reverse is true with Vc being small and Vr being large.

While the circuit above is that of an RC Low Pass Filter circuit, it can also be classed as a frequency variable potential divider circuit similar to the one we looked at in the Resistors tutorial. In that tutorial we used the following equation to calculate the output voltage for two single resistors connected in series.

$$V_{out} = V_{in} \times \frac{R_2}{R_1 + R_2}$$

where: $R_1 + R_2 = R_T$, the total resistance of the circuit

We also know that the capacitive reactance of a capacitor in an AC circuit is given as:

$$X_C = \frac{1}{2\pi f C} \text{ in Ohm's}$$

Opposition to current flow in an AC circuit is called impedance, symbol Z and for a series circuit consisting of a single resistor in series with a single capacitor, the circuit impedance is calculated as:

$$Z = \sqrt{R^2 + X_C^2}$$

Then by substituting our equation for impedance above into the resistive potential divider equation gives us:

<u>RC POTENTIAL DIVIDER EQUATION</u>

$$V_{out} = V_{in} \times \frac{X_C}{\sqrt{R^2 + X_C^2}} = V_{in} \frac{X_C}{Z}$$

So, by using the potential divider equation of two resistors in series and substituting for impedance we can calculate the output voltage of an RC Filter for any given frequency.

LOW PASS FILTER EXAMPLE

A Low Pass Filter circuit consisting of a resistor of 4k7Ω in series with a capacitor of 47nF is connected across a 10v sinusoidal supply. Calculate the output voltage (Vout) at a frequency of 100Hz and again at frequency of 10,000Hz or 10kHz.

Voltage Output at a Frequency of 100Hz.

$$X_C = \frac{1}{2\pi f C} = \frac{1}{2\pi \times 100 \times 47 \times 10^{-9}} = 33{,}863\Omega$$

$$V_{OUT} = V_{IN} \times \frac{X_C}{\sqrt{R^2 + X_C^2}} = 10 \times \frac{33863}{\sqrt{4700^2 + 33863^2}} = 9.9v$$

Voltage Output at a Frequency of 10,000Hz (10kHz).

$$X_C = \frac{1}{2\pi f C} = \frac{1}{2\pi \times 10{,}000 \times 47 \times 10^{-9}} = 338.6\Omega$$

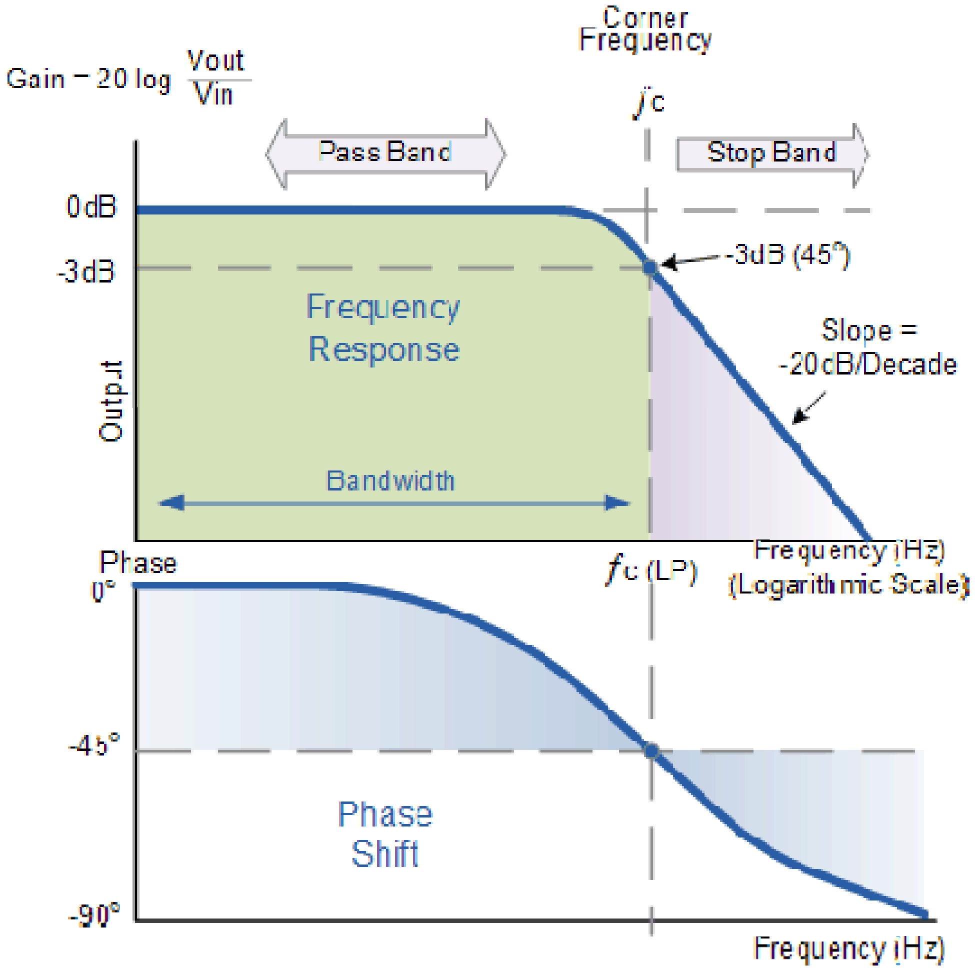
$$V_{OUT} = V_{IN} \times \frac{X_C}{\sqrt{R^2 + X_C^2}} = 10 \times \frac{338.6}{\sqrt{4700^2 + 338.6^2}} = 0.718v$$

FREQUENCY RESPONSE

We can see from the results above that as the frequency applied to the RC network increases from 100Hz to 10 kHz, the voltage dropped across the capacitor and therefore the output voltage (Vout) from the circuit decreases from 9.9v to 0.718v.

By plotting the networks output voltage against different values of input frequency, the Frequency Response Curve or Bode Plot function of the low pass filter circuit can be found, as shown below.

Frequency Response of a 1st-order Low Pass Filter



The Bode Plot shows the Frequency Response of the filter to be nearly flat for low frequencies and the entire input signal is passed directly to the output, resulting in a gain of nearly 1, called unity, until it reaches its Cut-off Frequency point (ƒc). This is because the reactance of the capacitor is high at low frequencies and blocks any current flow through the capacitor.

After this cut-off frequency point the response of the circuit decreases to zero at a slope of -20dB/ Decade or (-6dB/Octave) "roll-off". Note that the angle of the slope, this -20dB/ Decade roll-off will always be the same for any RC combination.

Any high frequency signals applied to the low pass filter circuit above this cut-off frequency point will become greatly attenuated, that is they rapidly decrease. This happens because at very high frequencies the reactance of the capacitor becomes so low that it gives the effect of a short circuit condition on the output terminals resulting in zero output.

Then by carefully selecting the correct resistor-capacitor combination, we can create a RC circuit that allows a range of frequencies below a certain value to pass through the circuit unaffected while any frequencies applied to the circuit above this cut-off point to be attenuated, creating what is commonly called a Low Pass Filter.

For this type of “Low Pass Filter” circuit, all the frequencies below this cut-off, ƒc point that are unaltered with little or no attenuation and are said to be in the filters Pass band zone. This pass band zone also represents the Bandwidth of the filter. Any signal frequencies above this point cut-off point are generally said to be in the filters Stop band zone and they will be greatly attenuated.

This “Cut-off”, “Corner” or “Breakpoint” frequency is defined as being the frequency point where the capacitive reactance and resistance are equal, R = Xc = 4k7Ω. When this occurs the output signal is attenuated to 70.7% of the input signal value or -3dB (20 log (Vout/Vin)) of the input. Although R = Xc, the output is not half of the input signal. This is because it is equal to the vector sum of the two and is therefore 0.707 of the input.

As the filter contains a capacitor, the Phase Angle (Φ) of the output signal LAGS behind that of the input and at the -3dB cut-off frequency (ƒc) and is -45º out of phase. This is due to the time taken to charge the plates of the capacitor as the input voltage changes, resulting in the output voltage (the voltage across the capacitor) “lagging” behind that of the input signal. The higher the input frequency applied to the filter the more the capacitor lags and the circuit becomes more and more “out of phase”.

The cut-off frequency point and phase shift angle can be found by using the following equation:

CUT-OFF FREQUENCY AND PHASE SHIFT

$$f c = \frac{1}{2\pi RC} = \frac{1}{2\pi \times 4700 \times 47 \times 10^{-9}} = 720\text{Hz}$$

$$\text{Phase Shift } \varphi = -\arctan(2\pi f RC)$$

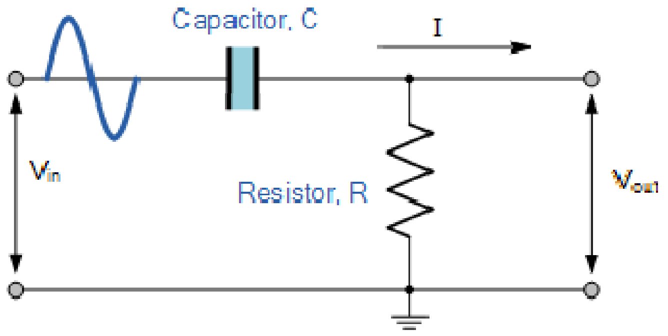
Then for our simple example of a “Low Pass Filter” circuit above, the cut-off frequency (ƒc) is given as720Hz with an output voltage of 70.7% of the input voltage value and a phase shift angle of -45º.

HIGH PASS FILTERS

A High Pass Filter or HPF, is the exact opposite to that of the previously seen Low Pass filter circuit, as now the two components have been interchanged with the output signal (Vout) being taken from across the resistor as shown.

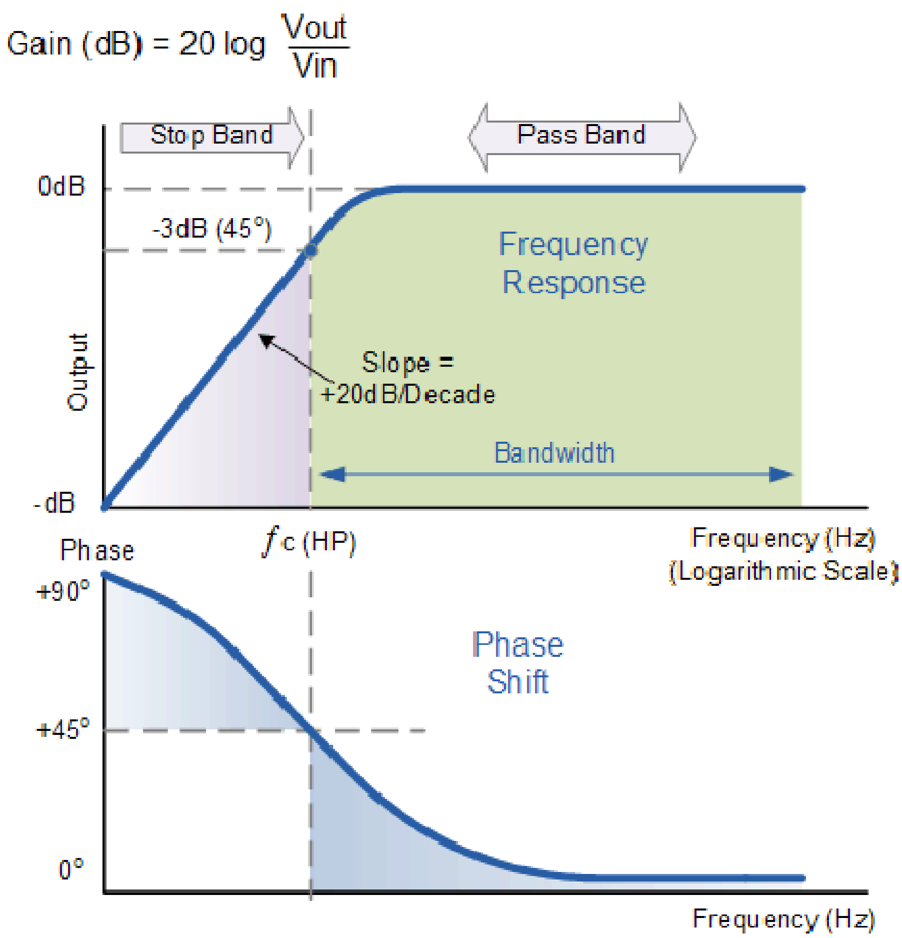
Where as the low pass filter only allowed signals to pass below its cut-off frequency point, ƒc, the passive high pass filter circuit as its name implies, only passes signals above the selected cut-off point, ƒc eliminating any low frequency signals from the waveform. Consider the circuit below.

THE HIGH PASS FILTER CIRCUIT



In this circuit arrangement, the reactance of the capacitor is very high at low frequencies so the capacitor acts like an open circuit and blocks any input signals at Vin until the cut-off frequency point (ƒc) is reached. Above this cut-off frequency point the reactance of the capacitor has reduced sufficiently as to now act more like a short circuit allowing the entire input signal to pass directly to the output as shown below in the High Pass Frequency Response Curve.

FREQUENCY RESPONSE OF A 1ST ORDER HIGH PASS FILTER.



The Bode Plot or Frequency Response Curve above for a High Pass filter is the exact opposite to that of a low pass filter. Here the signal is attenuated or damped at low frequencies with the output increasing at +20dB/Decade (6dB/Octave) until the frequency reaches the cut-off point (ƒc) where again R = Xc. It has a response curve that extends down from infinity to the cut-off frequency, where the output voltage amplitude is $1/\sqrt{2}$ = 70.7% of the input signal value or -3dB (20 log (Vout/Vin)) of the input value.

Also we can see that the phase angle (Φ) of the output signal LEADS that of the input and is equal to+45° at frequency ƒc. The frequency response curve for a high pass filter implies that the filter can pass all signals out to infinity. However in practice, the high pass filter response does not extend to infinity but is limited by the electrical characteristics of the components used.

The cut-off frequency point for a first order high pass filter can be found using the same equation as that of the low pass filter, but the equation for the phase shift is modified slightly to account for the positive phase angle as shown below.

CUT-OFF FREQUENCY AND PHASE SHIFT

$$fc = \frac{1}{2\pi RC}$$

$$Phase\ Shift\ \phi = \arctan \frac{1}{2\pi fRC}$$

The circuit gain, Av which is given as Vout/Vin (magnitude) and is calculated as:

$$A_V = \frac{V_{OUT}}{V_{IN}} = \frac{R}{\sqrt{R^2 + XC^2}} = \frac{R}{Z}$$

$$\text{at low } f\text{:}\quad Xc \to \infty,\quad Vout = 0$$
$$\text{at high } f\text{:}\quad Xc \to 0,\quad Vout = Vin$$

HIGH PASS FILTER EXAMPLE.

Calculate the cut-off or "breakpoint" frequency (ƒc) for a simple high pass filter consisting of an82pF capacitor connected in series with a 240kΩ resistor.

$$fc = \frac{1}{2\pi RC} = \frac{1}{2\pi \times 240{,}000 \times 82 \times 10^{-12}} = 8{,}087\ Hz\ or\ 8kHz$$

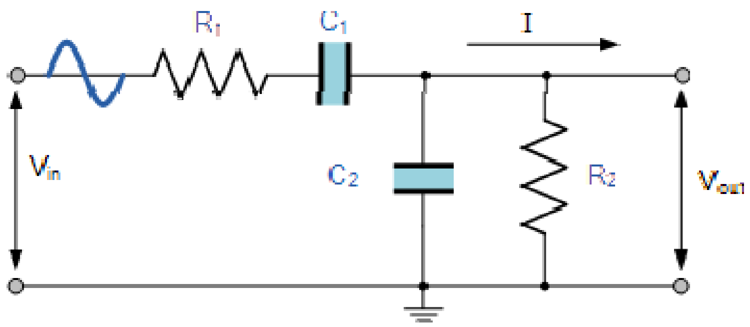
BAND PASS FILTERS

The cut-off frequency or ƒc point in a simple RC passive filter can be accurately controlled using just a single resistor in series with a non-polarized capacitor, and depending upon which way around they are connected either a low pass or a high pass filter is obtained.

One simple use for these types of Passive Filters is in audio amplifier applications or circuits such as in loudspeaker crossover filters or pre-amplifier tone controls. Sometimes it is necessary to only pass a certain range of frequencies that do not begin at 0Hz, (DC) or end at some high frequency point but are within a certain frequency band, either narrow or wide.

By connecting or "cascading" together a single Low Pass Filter circuit with a High Pass Filter circuit, we can produce another type of passive RC filter that passes a selected range or "band" of frequencies that can be either narrow or wide while attenuating all those outside of this range. This new type of passive filter arrangement produces a frequency selective filter known commonly as a Band Pass Filter or BPF for short.

BAND PASS FILTER CIRCUIT



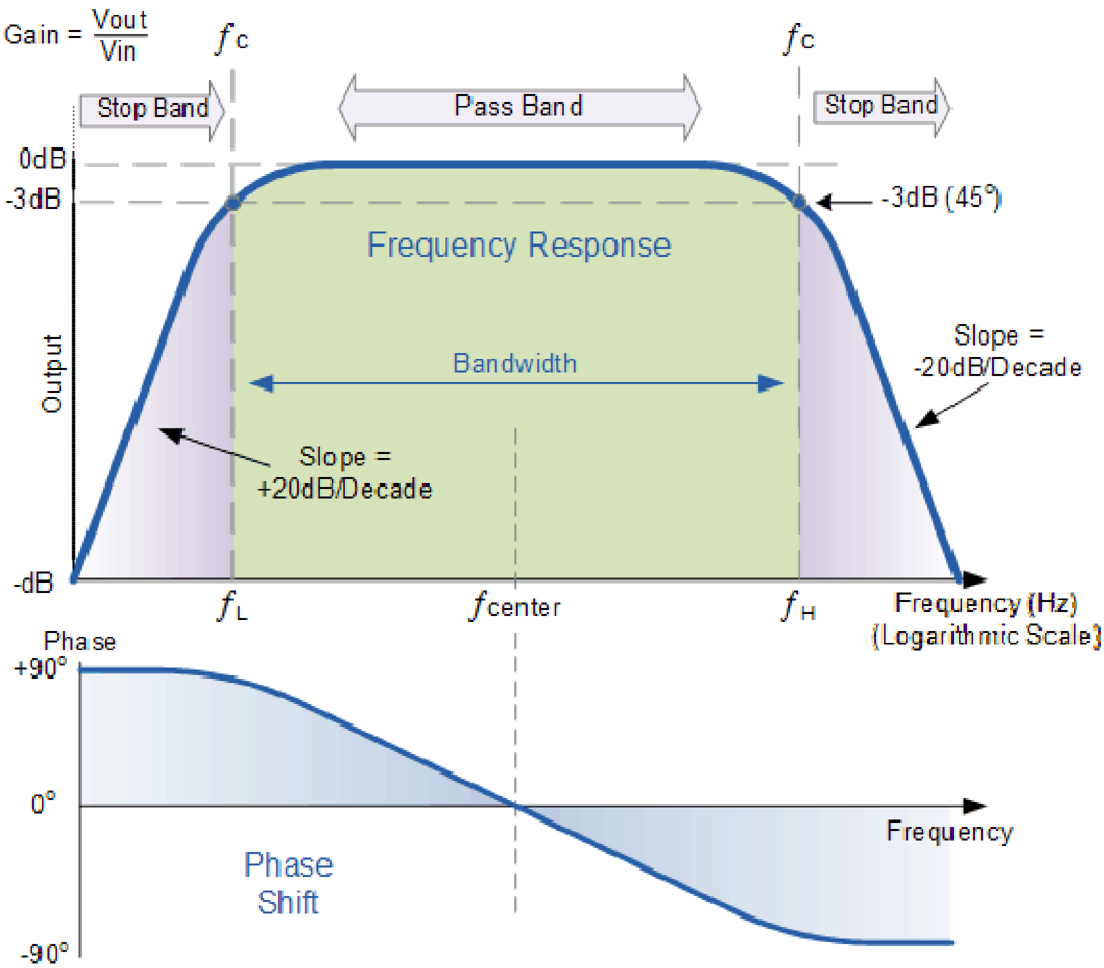
Unlike a low pass filter that only pass signals of a low frequency range or a high pass filter which pass signals of a higher frequency range, a Band Pass Filters passes signals within a certain "band" or "spread" of frequencies without distorting the input signal or introducing extra noise. This band of frequencies can be any width and is commonly known as the filters Bandwidth.

Bandwidth is commonly defined as the frequency range that exists between two specified frequency cut-off points (ƒc), that are 3dB below the maximum centre or resonant peak while attenuating or weakening the others outside of these two points.

Then for widely spread frequencies, we can simply define the term "bandwidth", BW as being the difference between the lower cut-off frequency ($ƒc_{LOWER}$) and the higher cut-off frequency ($ƒc_{HIGHER}$) points. In other words, BW = $ƒ_H – ƒ_L$. Clearly for a pass band filter to function correctly, the cut-off frequency of the low pass filter must be higher than the cut-off frequency for the high pass filter.

The "ideal" Band Pass Filter can also be used to isolate or filter out certain frequencies that lie within a particular band of frequencies, for example, noise cancellation. Band pass filters are known generally as second-order filters, (two-pole) because they have "two" reactive component, the capacitors, within their circuit design. One capacitor in the low pass circuit and another capacitor in the high pass circuit.

Frequency Response of a 2nd Order Band Pass Filter.



The Bode Plot or frequency response curve above shows the characteristics of the band pass filter. Here the signal is attenuated at low frequencies with the output increasing at a slope of +20dB/Decade (6dB/Octave) until the frequency reaches the "lower cut-off" point f_L. At this frequency the output voltage is again $1/\sqrt{2}$ = 70.7% of the input signal value or -3dB (20 log (Vout/Vin)) of the input.

The output continues at maximum gain until it reaches the "upper cut-off" point f_H where the output decreases at a rate of -20dB/Decade (6dB/Octave) attenuating any high frequency signals. The point of maximum output gain is generally the geometric mean of the two -3dB value between the lower and upper cut-off points and is called the "Centre Frequency" or "Resonant Peak" value fr. This geometric mean value is calculated as being $fr^2 = f_{(UPPER)} \times f_{(LOWER)}$.

A band pass filter is regarded as a second-order (two-pole) type filter because it has "two" reactive components within its circuit structure, then the phase angle will be twice that of the previously seen first-order filters, i.e., 180°. The phase angle of the output signal LEADS that of the input by $+90^{o}$ up to the centre or resonant frequency, ƒr point were it becomes "zero" degrees (0^{o}) or "in-phase" and then changes to LAG the input by -90^{o} as the output frequency increases.

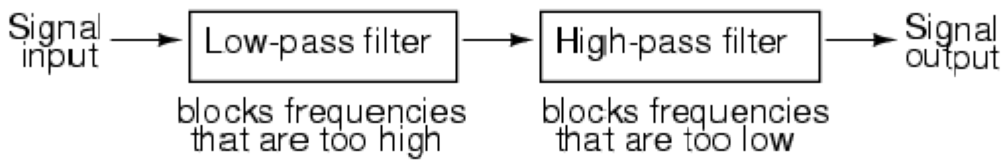
The upper and lower cut-off frequency points for a band pass filter can be found using the same formula as that for both the low and high pass filters, For example.

$$fc = \frac{1}{2\pi RC} Hz$$

Then clearly, the width of the pass band of the filter can be controlled by the positioning of the two cut-off frequency points of the two filters.

<u>Band Pass Filter Example</u>

A second-order band pass filter is to be constructed using RC components that will only allow a range of frequencies to pass above 1kHz (1,000Hz) and below 30kHz (30,000Hz). Assuming that both the resistors have values of 10kΩ´s, calculate the values of the two capacitors required.



<u>The High Pass Filter Stage</u>

The value of the capacitor C1 required to give a cut-off frequency $ƒ_L$ of 1kHz with a resistor value of10kΩ is calculated as:

$$C = \frac{1}{2\pi fc.R} = \frac{1}{2\pi x 1{,}000 x 10{,}000} = 15.8 nF$$

Then, the values of R1 and C1 required for the high pass stage to give a cut-off frequency of 1.0kHz are: R1 = 10kΩ´s and C1 = 15nF.

<u>The Low Pass Filter Stage</u>

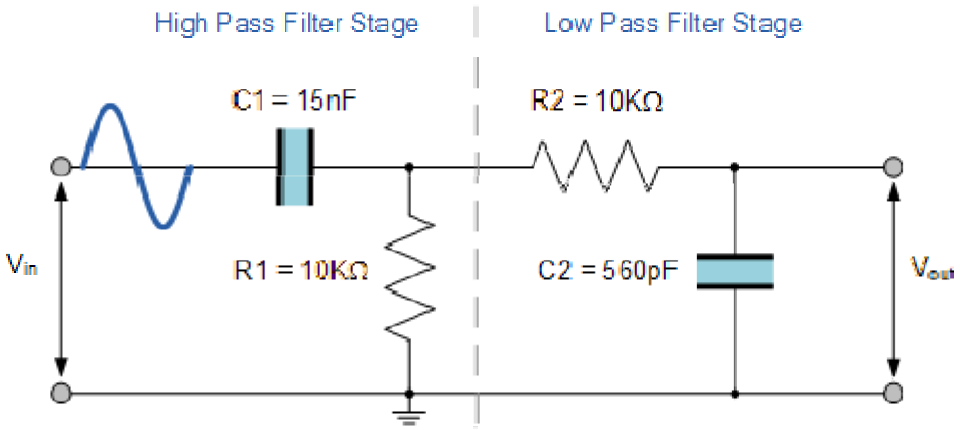
The value of the capacitor C2 required to give a cut-off frequency $ƒ_H$ of 30kHz with a resistor value of10kΩ is calculated as:

$$C = \frac{1}{2\pi fc.R} = \frac{1}{2\pi x 30{,}000 x 10{,}000} = 510 pF$$

Then, the values of R2 and C2 required for the low pass stage to give a cut-off frequency of 30kHz are, R = 10kΩ´s and C = 510pF. However, the nearest preferred value of the calculated capacitor value of 510pF is 560pF so this is used instead.

With the values of both the resistances R1 and R2 given as 10kΩ, and the two values of the capacitors C1 and C2 found for the high pass and low pass filters as 15nF and 560pF respectively, then the circuit for our simple passive Band Pass Filter is given as.

Completed Band Pass Filter Circuit



Band Pass Filter Resonant Frequency

We can also calculate the "Resonant" or "Centre Frequency" (ƒr) point of the band pass filter were the output gain is at its maximum or peak value. This peak value is not the arithmetic average of the upper and lower -3dB cut-off points as you might expect but is in fact the "geometric" or mean value. This geometric mean value is calculated as being $fr^2 = fc_{(UPPER)} \times fc_{(LOWER)}$ for example:

Centre Frequency Equation

$$fr = \sqrt{f_L \ x \ f_H}$$

- Where, f_r is the resonant or centre frequency
- f_L is the lower -3dB cut-off frequency point
- f_H is the upper -3db cut-off frequency point

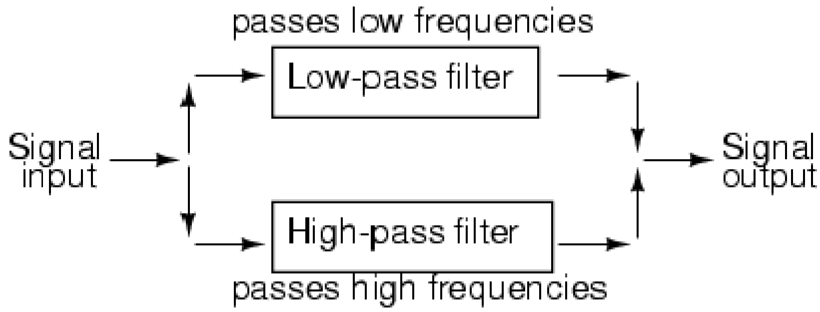
And in our simple example above, the calculated cut-off frequencies were found to be f_L = 1,060 Hz and f_H = 28,420 Hz using the filter values.

Then by substituting these values into the above equation gives a central resonant frequency of:

$$fr = \sqrt{f_L \ x \ f_H} = \sqrt{1{,}060 \ x \ 28{,}420} = 5{,}48 \, kHz$$

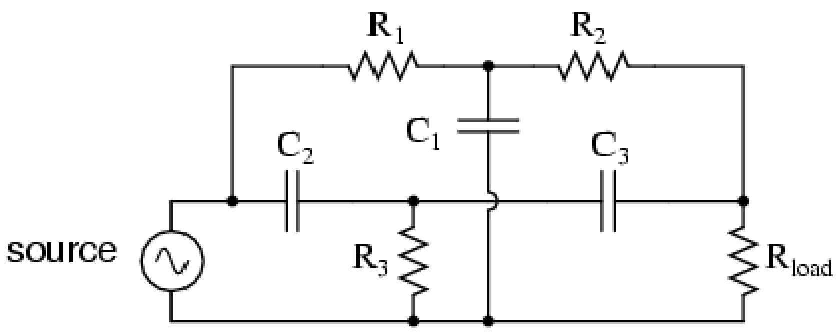
Band-stop filters

It is so called *band-elimination*, *band-reject*, or *notch* filters; this kind of filter passes all frequencies above and below a particular range settly the component values. Not surprisingly, it can be made out of a low-pass and a high-pass filter, just like the band-pass design, except that this time we connect the two filter sections in parallel with each other instead of in series. (Figure below)



System level block diagram of a band-stop filter.

Constructed using two capacitive filter sections, it looks something like (Figure below).

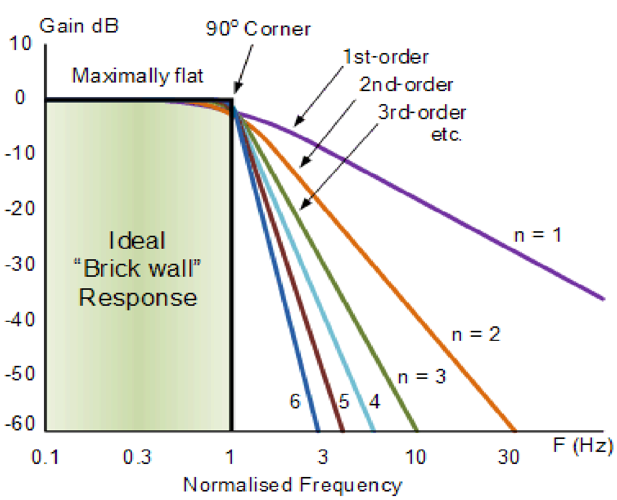


Butterworth Filter

A Butterworth filter is a type of signal processing filter designed to have a frequency response as flat as possible in the passband. Hence the Butterworth filter is also known as “maximally flat magnitude filter”. It was invented in 1930 by the British engineer and physicist Stephen Butterworth in his paper titled “*On the Theory of Filter Amplifiers*”. The frequency response of the Butterworth filter is flat in the passband (i.e. a bandpass filter) and roll-offs towards zero in the stopband. The rate of roll-off response depends on the order of the filter. The number of reactive elements used in the filter circuit will decide the order of the filter.
The inductor and capacitor are reactive elements used in filters. But in the case of Butterworth filter only capacitors are used. So, the number of capacitors will decide the order of the filter.
Here, we will discuss the Butterworth filter with a low pass filter. Similarly, the high pass filter can be designed by just changing the position of resistance and capacitance.

Butterworth Low Pass Filter Design

While designing the filter, the designer tries to achieve a response near to the ideal filter. It is very difficult to match results with the exact ideal characteristic. We need to use complex higher-order If you increase the order of the filter, the number of cascade stages with the filter is also increased. But in practice, we cannot achieve Butterworth’s ideal frequency response. Because it produces excessive ripple in the passband.In Butterworth filter, mathematically it is possible to get flat frequency response from 0 Hz to the cut-off frequency at -3dB with no ripple. If the frequency is more than the cut-off frequency, it will roll-off towards zero with the rate of -20 dB/decade for the first-order filter.If you increase the order of the filter, the rate of a roll-off period is also increased. And for second-order, it is -40 dB/decade. The quality factor for the Butterworth filter is 0.707.
The below figure shows the frequency response of the Butterworth filter for various orders of the filter



Frequency Response of Butterworth FilterThe generalized form of frequency response for nth-order Butterworth low-pass filter is;

$$H(j\omega) = \frac{1}{\sqrt{1 + \varepsilon^2 (\frac{\omega}{\omega_C})^{2n}}}$$

Where,

n = order of the filter,
ω = operating frequency (passband frequency) of circuit
ω_C = Cut-off frequency
ε = maximum passband gain = A_{max}

The below equation is used to find the value of ε.

$$H_1 = \frac{H_0}{\sqrt{1 + \varepsilon^2}}$$

Where,
H_1 = minimum passband gain
H_0 = maximum passband gain

First-order Lowpass Butterworth Filter

The lowpass filter is a filter that allows the signal with the frequency is lower than the cutoff frequency and attenuates the signals with the frequency is more than cutoff frequency. In the first-order filter, the number of reactive components is only one. The below figure shows the circuit diagram of the first-order lowpass Butterworth filter.

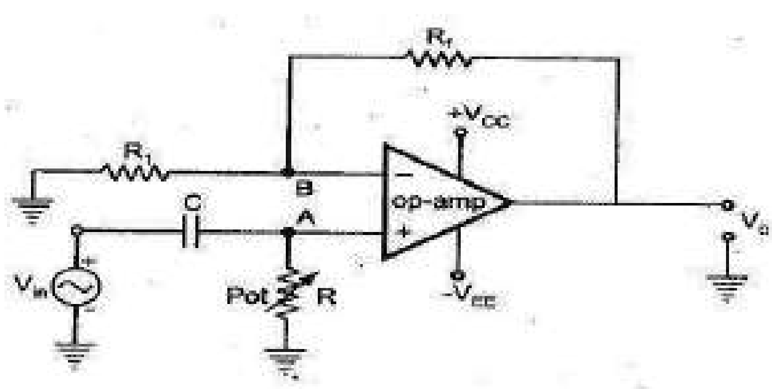


Fig. 2.79 First order high pass Butterworth filter

The low pass Butterworth filter is an active Low pass filter as it consists of the op-amp. This op-amp operates on non-inverting mode. Hence, the gain of the filter will decide by the resistor R_1 and R_F. And the cutoff frequency decides by R and C.
Now, if you apply the voltage divider rule at point Va and find the voltage across a capacitor. It is given as;

$$V_a = \frac{-jX_C}{R - jX_C} V_{in}$$

$$V_a = \frac{-j(\frac{1}{2\pi fC})}{R - j(\frac{1}{2\pi fC})} V_{in}$$

$$V_a = \frac{-j}{2\pi fRC - j} V_{in}$$

$$V_a = \frac{V_{in}}{1 - \frac{2\pi fRC}{j}}$$

$$V_a = \frac{V_{in}}{1 + j2\pi fRC}$$

Because of the non-inverting configuration of an op-amp,

$$V_0 = \left(1 + \frac{R_f}{R_1}\right) V_a$$

$$V_0 = \left(1 + \frac{R_f}{R_1}\right) \frac{V_{in}}{1 + j2\pi fRC}$$

$$\frac{V_0}{V_a} = \frac{A_f}{1 + j\frac{f}{f_c}}$$

WHERE

$$A_f = 1 + \frac{R_F}{R_1}$$

A_f = Gain of filter in Passband

$$f_c = \frac{1}{2\pi RC}$$

f_c = Cutoff Frequency

f = Operating Frequency

$$\frac{V_0}{V_a} = \left|\frac{V_0}{V_a}\right| \angle \phi$$

$$\left|\frac{V_0}{V_a}\right| = \frac{A_f}{\sqrt{1 + j\left(\frac{f}{f_c}\right)^2}}$$

$$\phi = -\tan^{-1}\left(\frac{f}{f_c}\right)$$

1. At very low frequency, $f << f_c$

$$\left|\frac{V_0}{V_a}\right| \approx A_f (Constant)$$

2. At cutoff frequency, $f = f_c$

$$\left|\frac{V_0}{V_a}\right| = \frac{A_f}{\sqrt{2}} = 0.707 A_f$$

3. At high frequency, $f > f_c$

$$\left|\frac{V_0}{V_a}\right| < A_f$$

The below figure shows the frequency response of first-order lowpass Butterworth filter.

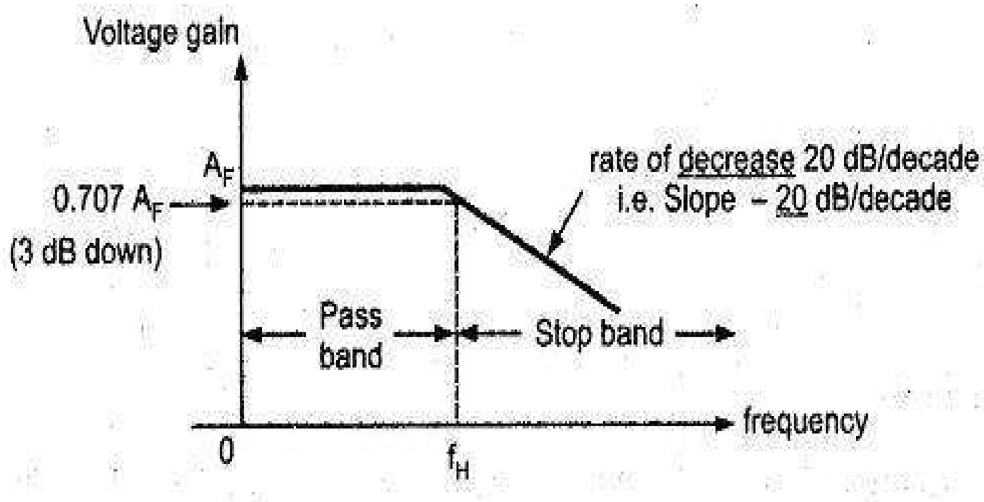


Fig. 2.75 Frequency response

Second-order Butterworth Filter

The second-order Butterworth filter consists of two reactive components. The circuit diagram of a second-order low pass Butterworth filter is as shown in the below figure.

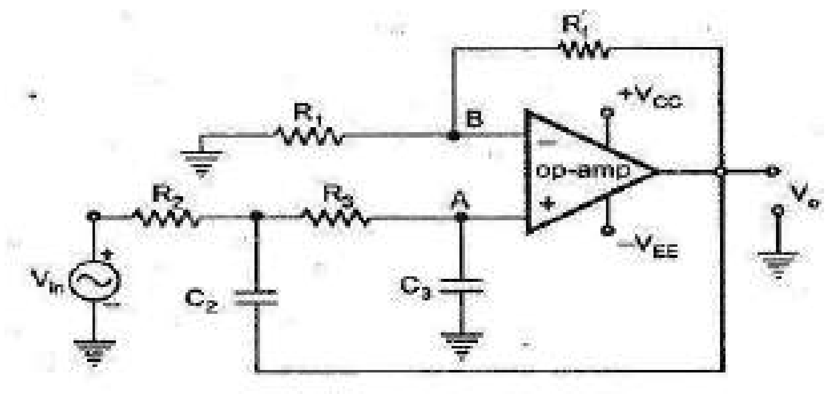


Fig. 2.76 Second order low pass butterworth filter

In this type of filter, resistor R and R_F are the negative feedback of op-amp. And the cutoff frequency of the filter decides by R_2, R_3, C_2, and C_3.The second-order lowpass Butterworth filter consists of two back-to-back connected RC networks. And R_L is the load resistance. First-order and second-order Butterworth filters are very important. Because we can get higher-order Butterworth filter by just cascading of the first-order and second-order Butterworth filters.

Let's analyse the circuit of second-order Butterworth filter,

Apply Kirchhoff's Current Law at point V_1.

$$I_1 = I_2 + I_3$$

$$\frac{V_{in} - V_1}{R_2} = \frac{V_1 - V_0}{\frac{1}{sC_2}} + \frac{V_1 - V_a}{R_3} \tag{1}$$

Using potential divider rule at point V_a

$$V_a = V_1 \left[\frac{\frac{1}{sC_3}}{R_3 + \frac{1}{sC_3}} \right]$$

$$V_a = V_1 \left[\frac{\frac{1}{sC_3}}{\frac{R_3 sC_3 + 1}{sC_3}} \right]$$

$$V_a = \frac{V_1}{1 + sR_3C_3}$$

Put the value of V_1 in equation-(1) $\quad V_1 = V_a(1 + sR_3C_3)$

$$\frac{V_{in} - V_a(1 + sR_3C_3)}{R_2} = \frac{V_a(1 + sR_3C_3) - V_0}{\frac{1}{sC_2}} + \frac{V_a(1 + sR_3C_3) - V_a}{\frac{1}{R_3}}$$

$$\frac{V_{in}}{R_2} - \frac{V_a(1 + sR_3C_3)}{R_2} = \frac{V_a(1 + sR_3C_3)}{\frac{1}{sC_2}} - \frac{V_0}{\frac{1}{sC_2}} + \frac{V_a(1 + sR_3C_3)}{R_3} - \frac{V_a}{R_3}$$

$$\frac{V_{in}}{R_2} + \frac{V_0}{\frac{1}{sC_2}} = \frac{V_a(1 + sR_3C_3)}{\frac{1}{sC_2}} + \frac{V_a(1 + sR_3C_3)}{R_2} + \frac{V_a(1 + sR_3C_3)}{R_3} - \frac{V_a}{R_3}$$

$$\frac{V_{in}}{R_2} + V_0sC_2 = V_a\left[sC_2(1 + sR_3C_3) + \frac{(1 + sR_3C_3)}{R_2} + \frac{(1 + sR_3C_3)}{R_3} - \frac{1}{R_3}\right]$$

$$\frac{V_{in} + V_0sC_2R_2}{R_2} = V_a\left[\frac{R_3R_2sC_2(1 + sR_3C_3) + R_3(1 + sR_3C_3) + R_2(1 + sR_3C_3) - R_2}{R_2R_3}\right]$$

$$R_3\left(V_{in} + V_0sC_2R_2\right) = V_a\left[R_3R_2sC_2(1 + sR_3C_3) + R_3(1 + sR_3C_3) + R_2(1 + sR_3C_3) - R_2\right]$$

$$R_3V_{in} + V_0sC_2R_2R_3 = V_a\left[(1 + sR_3C_3)\left(R_3R_2sC_2 + R_3 + R_2\right) - R_2\right]$$

$$V_a = \frac{R_3V_{in} + V_0sC_2R_2R_3}{(1 + sR_3C_3)\left(R_3R_2sC_2 + R_3 + R_2\right) - R_2}$$

Because of the non-inverting configuration of an op-amp,

$$V_0 = A_fV_a$$

Where,

$$A_f = 1 + \frac{R_f}{R_1} = Gain\ of\ filter\ in\ passband$$

$$V_0 = A_f\left[\frac{R_3V_{in} + V_0sC_2R_2R_3}{(1 + sR_3C_3)\left(R_3R_2sC_2 + R_3 + R_2\right) - R_2}\right]$$

$$V_0 - \frac{A_fV_0sC_2R_2R_3}{(1 + sR_3C_3)\left(R_3R_2sC_2 + R_3 + R_2\right) - R_2} = \frac{A_fR_3V_{in}}{(1 + sR_3C_3)\left(R_3R_2sC_2 + R_3 + R_2\right) - R_2}$$

$$V_0\left[(1 + sR_3C_3)(R_3R_2sC_2 + R_3 + R_2) - R_2 - A_fsC_2R_2R_3\right] = A_fR_3V_{in}$$

$$\frac{V_0}{V_{in}} = \frac{A_fR_3}{\left[(1 + sR_3C_3)(R_3R_2sC_2 + R_3 + R_2) - R_2 - A_fsC_2R_2R_3\right]}$$

Rearrange this equation,

$$\frac{V_0}{V_{in}} = \frac{A_f R_3}{[(1 + sR_3C_3)(R_2 + R_3 + sR_2R_3C_2) - R_2 - sA_f R_2R_3C_2]}$$

$$\frac{V_0}{V_{in}} = \frac{A_f R_3}{[(R_2 + R_3 + sR_2R_3C_2 + sR_2R_3C_3 + sR_3^2C_3 + s^2R_2R_3^2C_2C_3) - R_2 - sA_f R_2R_3C_2]}$$

$$\frac{V_0}{V_{in}} = \frac{A_f R_3}{s^2R_2R_3^2C_2C_3 + s(R_2R_3C_2 + R_2R_3C_3 + R_3^2C_3 - A_f R_2R_3C_2) + R_3}$$

$$\frac{V_0}{V_{in}} = \frac{A_f R_3}{R_2R_3^2C_2C_3\left(s^2 + s\frac{R_2R_3C_2 + R_2R_3C_3 + R_3^2C_3 - A_f R_2R_3C_2}{R_2R_3^2C_2C_3} + \frac{R_3}{R_2R_3^2C_2C_3}\right)}$$

$$\frac{V_0}{V_{in}} = \frac{A_f}{R_2R_3C_2C_3\left(s^2 + s\frac{R_2C_2 + R_3C_3 + R_3C_3 - A_f R_2C_2}{R_2R_3C_2C_3} + \frac{1}{R_2R_3C_2C_3}\right)}$$

$$\frac{V_0}{V_{in}} = \frac{\frac{A_f}{R_2R_3C_2C_3}}{\left(s^2 + s\frac{R_2C_2 + R_3C_3 + R_3C_3 - A_f R_2C_2}{R_2R_3C_2C_3} + \frac{1}{R_2R_3C_2C_3}\right)}$$

Compare this equation with the standard form transfer function for second-order Butterworth filter. And that is,

$$\frac{V_0}{V_{in}} = \frac{A}{s^2 + 2\zeta\omega_c s + \omega_c^2}$$

By comparing above equations, we can find the equation of cutoff frequency and overall gain for the second-order lowpass Butterworth filter.

The gain of filter is,

$$A_{max} = \frac{A_f}{R_2 R_3 C_2 C_3}$$

And the Cutoff frequency of filter is ,

$$\omega_c^2 = \frac{1}{R_2 R_3 C_2 C_3}$$

$$f_c = \frac{1}{2\pi\sqrt{R_2 R_3 C_2 C_3}}$$

Now, if we consider the value of R_2 is same as R_3 and the value of C_2 is same as C_3.

$$R_2 = R_3 = R \quad and \quad C_2 = C_3 = C$$

$$f_c = \frac{1}{2\pi RC}$$

Now if we put above values in transfer function,

$$\frac{V_0}{V_{in}} = \frac{\frac{A_f}{R^2C^2}}{s^2 + s\frac{RC+RC+RC-A_f RC}{R^2C^2} + \frac{1}{R^2C^2}}$$

$$\omega_c = \frac{1}{RC}$$

$$\frac{V_0}{V_{in}} = \frac{A_f \omega^2}{s^2 + s(3 - A_f)\omega + \omega^2}$$

From above equation, the quality factor Q is equal to,

$$Q = \frac{1}{3 - A_f}$$

We can say that, the quality factor is only depends on the gain of filter. And the value of gain should not more than 3. If the value of gain is more than 3, the system will be unstable.

The value of quality factor is 0.707 for the Butterworth filter. And if we put this value in equation of quality factor, we can find the value of gain.

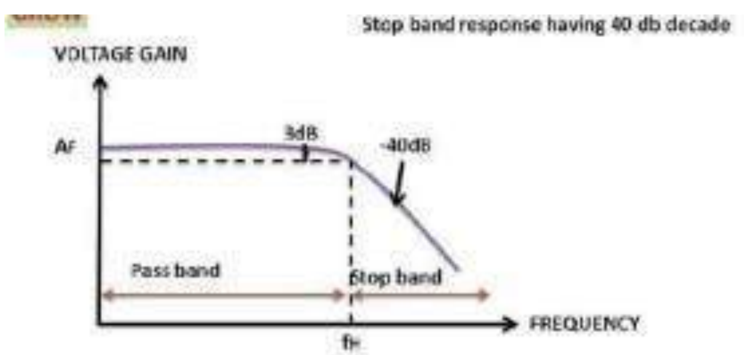
$$0.707 = \frac{1}{3 - A_f}$$

$$A_f = 1.586$$

$$1 + R_f R_1 = 1.586$$

$$R_f R_1 = 0.586$$

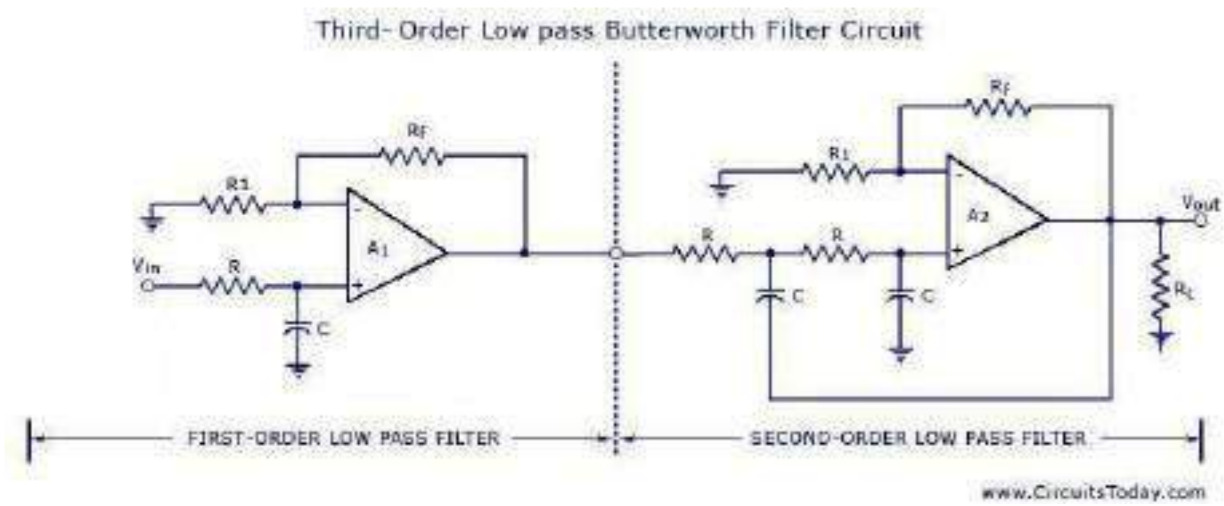
While designing the second-order Butterworth filter above relation must be satisfy. The frequency response of this filter is as shown in below figure.



<u>Third-order Lowpass Butterworth Filter</u>

Third-order lowpass Butterworth filter can design by cascading the first-order and second-order Butterworth filter.

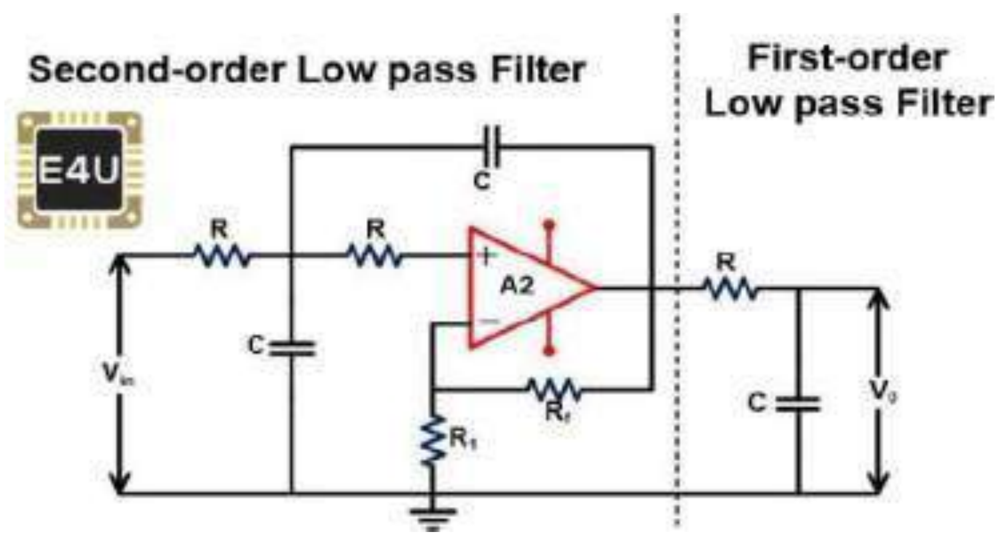
The below figure shows the circuit diagram of the third-order lowpass Butterworth filter.



Third-order Low Pass Butterworth Filter

In this figure, the first part shows the first-order lowpass Butterworth filter, and the second part shows the second-order lowpass Butterworth filter.

But in this condition, the voltage gain of the first part is optional and it can be set at any value. Therefore, the first op-amp is not taking part in voltage gain. Hence, the figure for the third-order low pass filter can be expressed as below figure also;

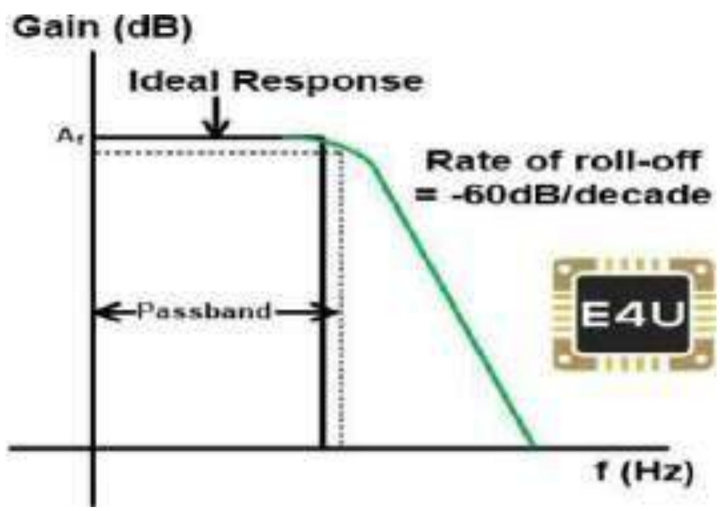


;

The voltage gain of a second-order filter affects the flatness of frequency response. If the gain of the second-order filter is kept at 1.586, the gain will down 3db for each part. So, the overall gain will down 6dB at the cutoff frequency.

By increasing the voltage gain of the second-order filter, we can offset the cumulative loss of voltage gain.

In the third-order Butterworth filter, the rate of a roll-off period is -60dB/decade. And the frequency response of this filter is nearer to the ideal Butterworth filter compared to the first and second-order filters. The frequ



(frequency response of this filter is as shown in the below figure.)

<u>Fourth-order Lowpass Butterworth Filter</u>

Fourth-order Butterworth filter is established by the cascade connection of two second-order low pass Butterworth filters. The circuit diagram of the fourth-order lowpass Butterworth filter is as shown in the below figure.

<u>**COURSE CONTENTS**</u>

1. FUNDAMENTAL OF CONTROL SYSTEM
1.1. Classification of Control system
1.2. Open loop system & Closed loop system and its comparison
1.3. Effects of Feed back
1.4. Standard test Signals(Step, Ramp, Parabolic, Impulse Functions)
1.5. Servomechanism
2. MATHEMATICAL MODEL OF A SYSTEM
2.1. Transfer Function & Impulse response,
2.2. Properties, Advantages & Disadvantages of Transfer Function
2.3. Poles & Zeroes of transfer Function
2.4. Simple problems of transfer function of network.
2.5. Mathematical modeling of Electrical Systems(R, L, C, Analogous systems)
3. CONTROL SYSTEM COMPONENTS
3.1. Components of Control System
3.2. Gyroscope, Synchros, Tachometer, DC servomotors, Ac Servomotors.
4. BLOCK DIAGRAM ALGEBRA & SIGNAL FLOW GRAPHS
4.1. Definition: Basic Elements of Block Diagram
4.2. Canonical Form of Closed loop Systems
4.3. Rules for Block diagram reduction
4.4. Procedure for of Reduction of Block Diagram
4.5. Simple Problem for equivalent transfer function
4.6. Basic Definition in Signal Flow Graph & properties
4.7. Construction of Signal Flow graph from Block diagram
4.8. Mason‘s Gain formula
4.9. Simple problems in Signal flow graph for network
5. TIME RESPONSE ANALYSIS.
5 . 1 Time response of control system.
5 . 2 Standard Test signal.
5.2.1. Step signal,
5.2.2. Ramp Signal
5.2.3. Parabolic Signal
5.2.4. Impulse Signal
5 . 3 Time Response of first order system with:
5.3.1. Unit step response
5.3.2. Unit impulse response.
5 . 4 Time response of second order system to the unit step input.
5.4.1. Time response specification.
5.4.2. Derivation of expression for rise time, peak time, peak overshoot, settling time and steady state error.
5.4.3. Steady state error and error constants.
5 . 5 Types of control system.[Steady state errors in Type-0, Type-1, Type-2 system]
5 . 6 Effect of adding poles and zero to transfer function.
5 . 7 Response with P, PI, PD and PID controller.
6. ANALYSIS OF STABILITY BY ROOT LOCUS TECHNIQUE.
6 . 1 Root locus concept.
6 . 2 Construction of root loci.
6 . 3 Rules for construction of the root locus.
6 . 4 Effect of adding poles and zeros to G(s) and H(s).
7. FREQUENCY RESPONSE ANALYSIS.
7 . 1 Correlation between time response and frequency response.
7 . 2 Polar plots.
7 . 3 Bode plots.

7 . 4 All pass and minimum phase system.
7 . 5 Computation of Gain margin and phase margin.
7 . 6 Log magnitude versus phase plot.
7 . 7 Closed loop frequency response.

8. NYQUIST PLOT

8.1 Principle of argument.
8.2 Nyquist stability criterion.
8.3 Niquist stability criterion applied to inverse polar plot.
8.4 Effect of addition of poles and zeros to G(S) H(S) on the shape of Niquist plot.
8.5 Assessment of relative stability.
8.6 Constant M and N circle
8.7 Nicholas chat

COURSE OUT COME
After the completion of the course the students will be able to-
CO1- Acquire the basic concept of control system, pole and zero and can analyze system stability.
CO2- Implement mathematical tools to analyze a complete system.
CO3- Understand different time domain specification parameters & thus can apply that knowledge to analyze system stability in time domain using root locus technique.
CO4- Analyze systems absolute, relative and local stability using different frequency domain methods such as polar plot, Nyquist plot , Bode plot and Nicholas chart
CO5- Identify the needs of different types of controllers, Develop and analyze state space models.

Course Content

1. SIGNAL FLOW GRAPH.

1. Review of block diagrams and transfer functions of multivariable systems.
2. Construction of signal flow graph.
3. Basic properties of signal flow graph.
4. Signal flow graph algebra.
5. Construction of signal flow graph for control system.

Learning Resources:

Sl.No	Name of Authors	Title of the Book	Name of Publisher
1.	A. Ananda Kumar	Control System	PHI
2.	K. Padmanavan	Control System	IK
3.	I. J. Nagarath, M. Gopal	Control system Engineering	WEN
4.	A Natrajan,Ramesh Babu	Control system Engineering	Scientific
5.	D N Manik	Control Systems	Cengage

TRANSFER FUNCTION

Introduction:

Transfer function of a linear time-invariant system is defined as the ratio of the Laplace transform of output variable to the Laplace transform of input variable assuming all the initial conditions to be zero. The figure 1a shows the system in time domain whereas figure 1b shows the system in Laplace domain.

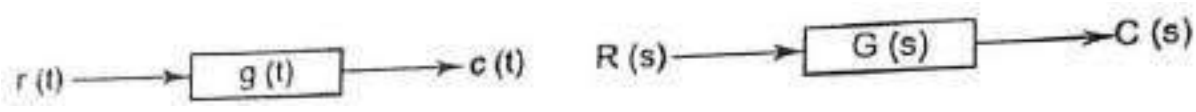


Figure 1a. system in time domain Fig1b. system in Laplace domain.

Figure 1. Transfer Function of a system

If G(s) be the transfer function of the system, we can write mathematically as

$$G(s) = \frac{\text{Laplace transform of output}}{\text{Laplace transform of input}} \text{ (all initial conditions are zero)}$$

$$= \frac{C(s)}{R(s)} \text{ (all initial conditions are zero)} \quad \text{.......(1)}$$

Example: Determine the transfer function of figure 2 shown below. $V_i(t)$ is the input to the system and $V_o(t)$ is the output of the system.

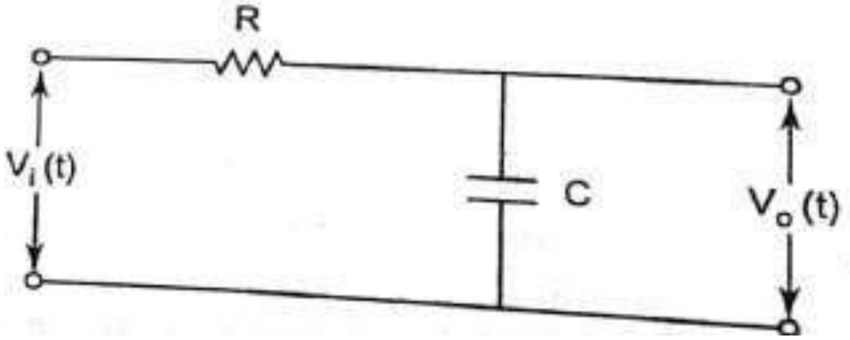


Figure 2

Solution- Let i(t) be the current flowing through the circuit using KVL we can write

$$V_i(t) = Ri(t) + \frac{1}{C}\int_{-\infty}^{t} i(t)dt$$

$$\text{And } V_o(t) = \frac{1}{C}\int_{-\infty}^{t} i(t)dt$$

Taking Laplace transfer of the above equation by assuming zero initial condition, we get

$$V_i(s) = RI(s) + \frac{1}{sC} I(s)$$

$$\text{And } V_o(s) = \frac{1}{sC} I(s)$$

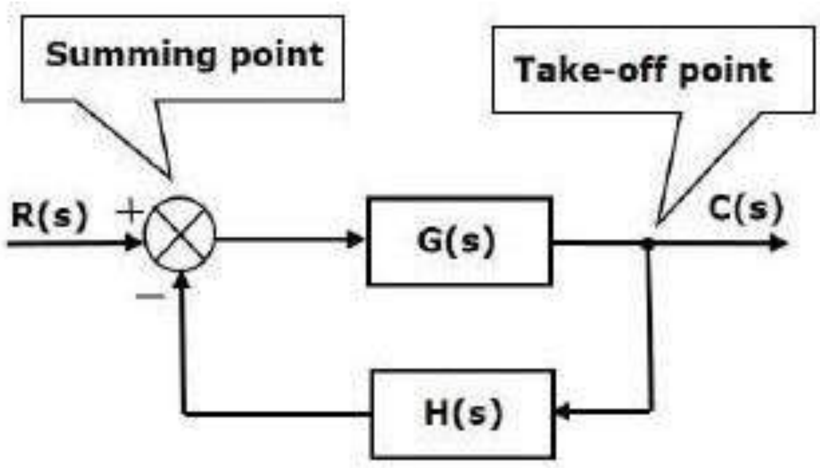
$$\therefore Transfer\ Function = (s) = \frac{Vo(s)}{Vi(s)} = \frac{1}{1+sCR}$$

BLOCK DIAGRAMS

Block diagram is the pictorial representation of system. It consists of a single block or a combination of blocks. Each block is a functional block.

Basic Elements of Block Diagram

The basic elements of a block diagram are a block, the summing point and the take-off point. Let us consider the block diagram of a closed loop control system as shown in the following figure to identify these elements.

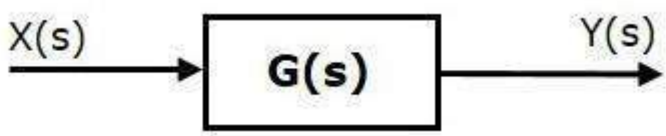


The above block diagram consists of two blocks having transfer functions G(s) and H(s). It is also having one summing point and one take-off point. Arrows indicate the direction of the flow of signals. Let us now discuss these elements one by one.

Block

The transfer function of a component is represented by a block. Block has single input and single output.

The following figure shows a block having input X(s), output Y(s) and the transfer function G(s).



Transfer Function, $G(s)=Y(s)/X(s)$

$$\Rightarrow Y(s)=G(s)X(s)$$

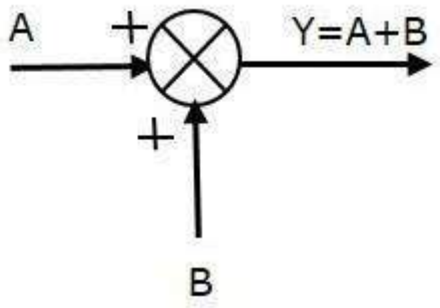
Output of the block is obtained by multiplying transfer function of the block with input.

Summing Point

The summing point is represented with a circle having cross (X) inside it. It has two or more inputs and single output. It produces the algebraic sum of the inputs. It also performs the summation or subtraction or combination of summation and subtraction of the inputs based on the polarity of the inputs. Let us see these three operations one by one.

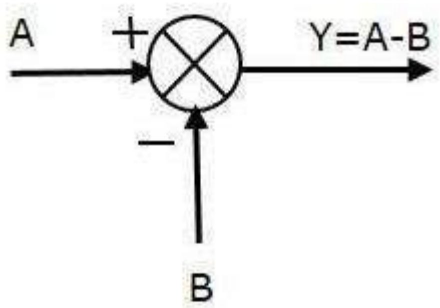
The following figure shows the summing point with two inputs (A, B) and one output (Y). Here, the inputs A and B have a positive sign. So, the summing point produces the output, Y as **sum of A and B.**

i.e.,Y = A + B.



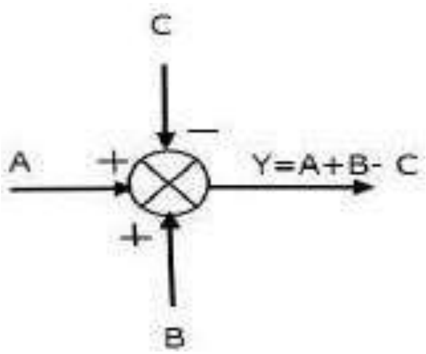
The following figure shows the summing point with two inputs (A, B) and one output (Y). Here, the inputs A and B are having opposite signs, i.e., A is having positive sign and B is having negative sign. So, the summing point produces the output **Y** as the **difference of A and B**.

Y = A + (-B) = A - B.



The following figure shows the summing point with three inputs (A, B, C) and one output (Y). Here, the inputs A and B are having positive signs and C is having a negative sign. So, the summing point produces the output **Y** as

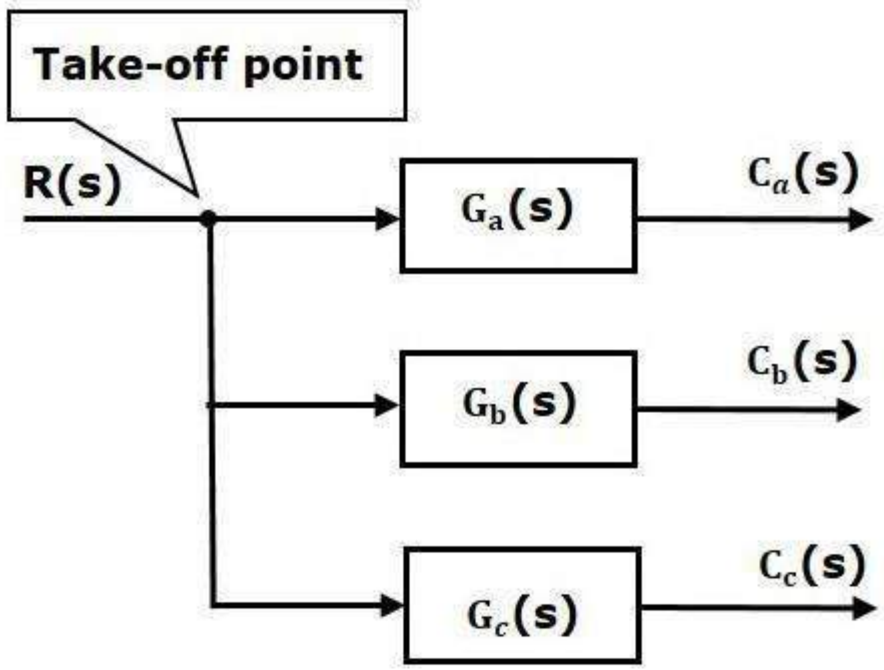
Y = A + B + (−C) = A + B − C.



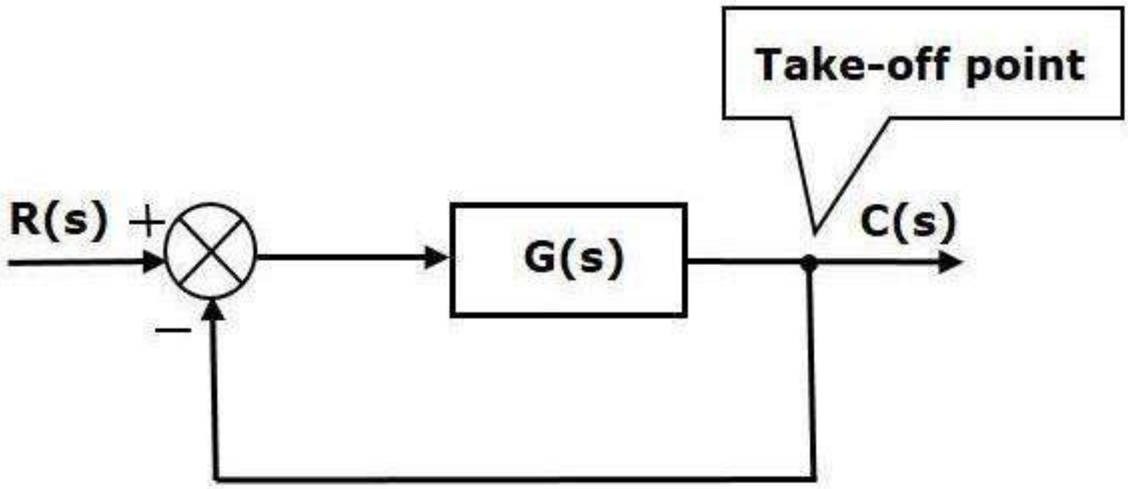
Take-off Point

The take-off point is a point from which the same input signal can be passed through more than one branch. That means with the help of take-off point, we can apply the same input to one or more blocks, summing points.

In the following figure, the take-off point is used to connect the same input, R(s) to two more blocks.



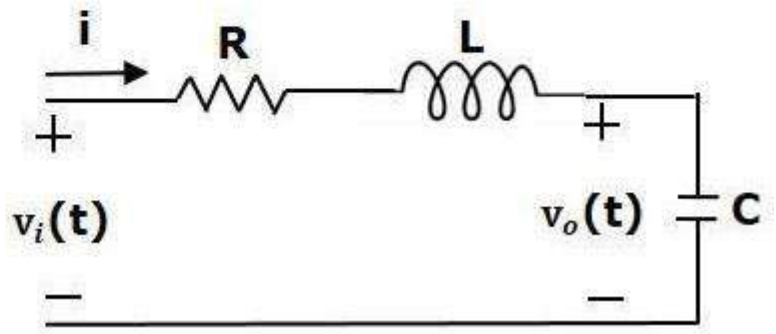
In the following figure, the take-off point is used to connect the output C(s), as one of the inputs to the summing point.



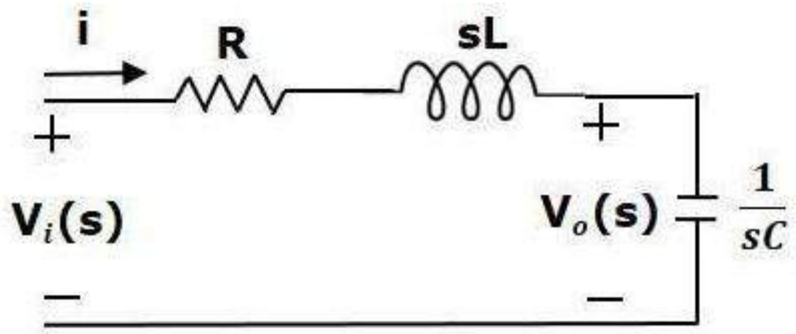
Block Diagram Representation of Electrical Systems

In this section, let us represent an electrical system with a block diagram. Electrical systems contain mainly three basic elements — **resistor, inductor and capacitor**.

Consider a series of RLC circuit as shown in the following figure. Where, $V_i(t)$ and $V_o(t)$ are the input and output voltages. Let i(t) be the current passing through the circuit. This circuit is in time domain.



By applying the Laplace transform to this circuit, will get the circuit in s-domain. The circuit is as shown in the following figure.



From the above circuit, we can write

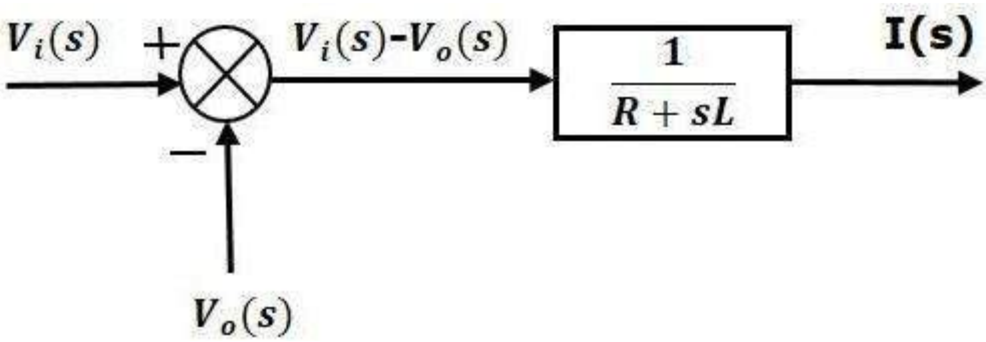
$$I(s)=[Vi(s)-Vo(s)]/R+sL$$

$\Rightarrow I(s)=\{1/R+sL\}\{Vi(s)-Vo(s)\}$ **(Equation 1)**

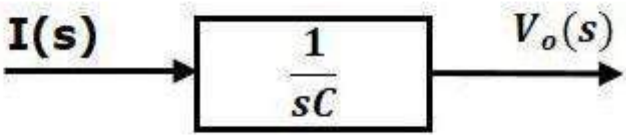
$Vo(s)=(1/sC)I(s)$ **(Equation 2)**

Let us now draw the block diagrams for these two equations individually. And then combine those block diagrams properly in order to get the overall block diagram of series of RLC Circuit (s-domain).

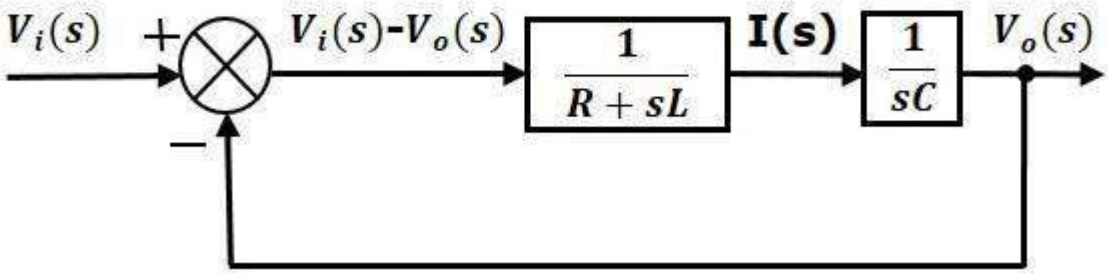
Equation 1 can be implemented with a block having the transfer function, 1/R+sL. The input and output of this block are {Vi(s)−Vo(s)} and I(s). We require a summing point to get {Vi(s)−Vo(s)}. The block diagram of Equation 1 is shown in the following figure.



Equation 2 can be implemented with a block having transfer function, 1/sC. The input and output of this block are I(s) and Vo(s). The block diagram of Equation 2 is shown in the following figure.



The overall block diagram of the series of RLC Circuit (s-domain) is shown in the following figure.



Similarly, you can draw the **block diagram** of any electrical circuit or system just by following this simple procedure.

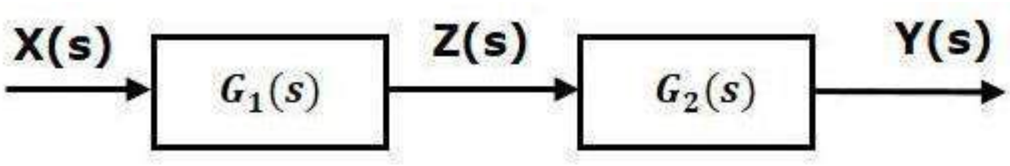
- Convert the time domain electrical circuit into an s-domain electrical circuit by applying Laplace transform.
- Write down the equations for the current passing through all series branch elements and voltage across all shunt branches.
- Draw the block diagrams for all the above equations individually.
- Combine all these block diagrams properly in order to get the overall block diagram of the electrical circuit (s-domain).

Block diagram reduction rules:

There are three basic types of connections between two blocks.

Rule 1: Series Connection

Series connection is also called **cascade connection**. In the following figure, two blocks having transfer functions G1(s) and G2(s) are connected in series.



For this combination, we will get the output Y(s) as

$$Y(s) = G2(s)\ Z(s)$$

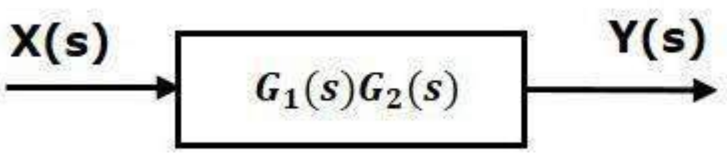
Where, Z(s) = G1(s) X(s)

$$\Rightarrow Y(s) = G2(s)\ [G1(s)\ X(s)] = G1(s)\ G2(s)\ X(s)$$

$$\Rightarrow Y(s) = \{G1(s)\ G2(s)\}\ X(s)$$

Compare this equation with the standard form of the output equation, Y(s) = G(s) X(s). Where, G(s) = G1(s) G2(s).

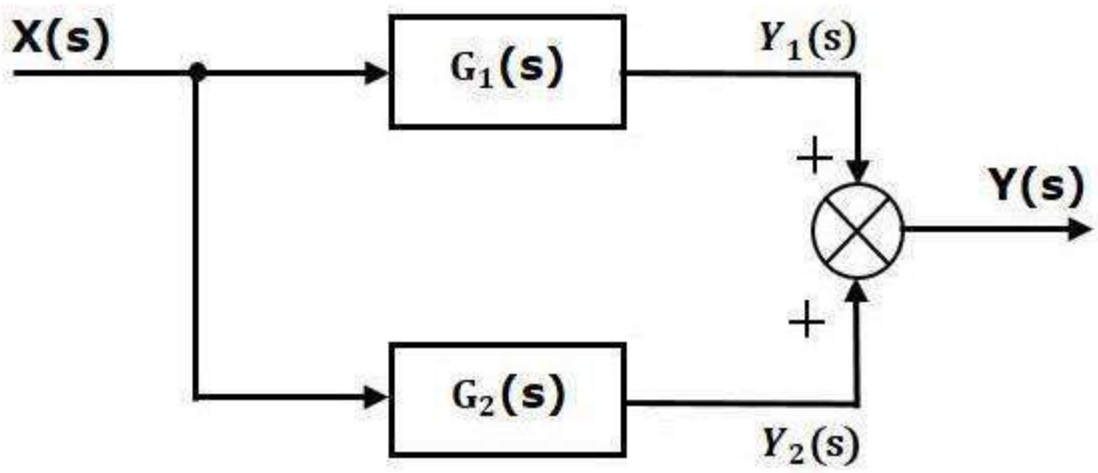
That means we can represent the **series connection** of two blocks with a single block. The transfer function of this single block is the **product of the transfer functions** of those two blocks. The equivalent block diagram is shown below.



Similarly, you can represent series connection of ‘n’ blocks with a single block. The transfer function of this single block is the product of the transfer functions of all those ‘n’ blocks.

Rule 2: Parallel Connection

The blocks which are connected in **parallel** will have the **same input**. In the following figure, two blocks having transfer functions G1(s) and G2(s) are connected in parallel. The outputs of these two blocks are connected to the summing point.



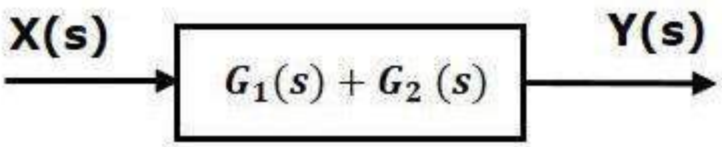
For this combination, we will get the output Y(s) as

$$Y(s) = Y1(s) + Y2(s)$$

Where, Y1(s) = G1(s) X(s)) and Y2(s) = G2(s) X(s)

$$\Rightarrow Y(s) = G1(s)\ X(s) + G2(s)\ X(s) = \{G1(s) + G2(s)\}\ X(s)$$

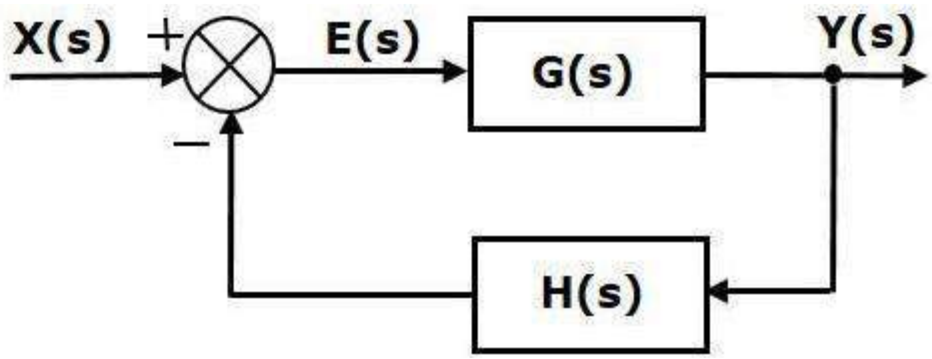
Compare this equation with the standard form of the output equation,
Y(s) = G(s) X(s)
Where, G(s) = G1(s) + G2(s)
That means we can represent the **parallel connection** of two blocks with a single block. The transfer function of this single block is the **sum of the transfer functions** of those two blocks. The equivalent block diagram is shown below.



Similarly, you can represent parallel connection of ‘n’ blocks with a single block. The transfer function of this single block is the algebraic sum of the transfer functions of all those ‘n’ blocks.

Rule 3: Feedback Connection

As we discussed in previous chapters, there are two types of **feedback** — positive feedback and negative feedback. The following figure shows negative feedback control system. Here, two blocks having transfer functions G(s) and H(s) form a closed loop.



The output of the summing point is -

$$E(s) = X(s) - H(s)\,Y(s)$$

The output Y(s) is -

$$Y(s) = E(s)\,G(s)$$

Substitute E(s) value in the above equation.

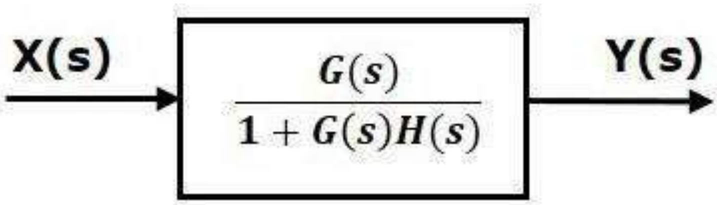
$$Y(s) = \{X(s) - H(s)Y(s)\}\,G(s)\}$$

$$Y(s)\,\{1 + G(s)\,H(s)\} = X(s)\,G(s)$$

$$\Rightarrow Y(s)/X(s) = G(s) / [1 + G(s)\,H(s)]$$

Therefore, the negative feedback closed loop transfer function is G(s) / [1+G(s) H(s)]

This means we can represent the negative feedback connection of two blocks with a single block. The transfer function of this single block is the closed loop transfer function of the negative feedback. The equivalent block diagram is shown below.



Similarly, you can represent the positive feedback connection of two blocks with a single block. The transfer function of this single block is the closed loop transfer function of the positive feedback, i.e., G(s) / [1 − G(s) H(s)]

Rule 4: Block Diagram Algebra for Summing Points

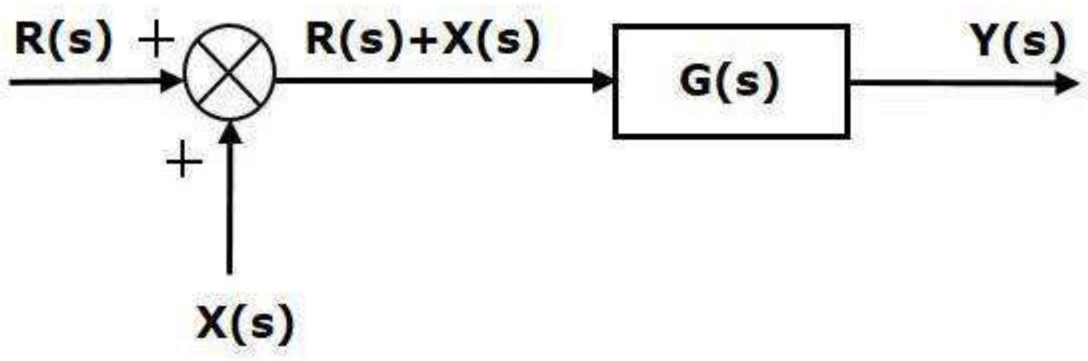
There are two possibilities of shifting summing points with respect to blocks −

- Shifting summing point after the block
- Shifting summing point before the block

Let us now see what kind of arrangements need to be done in the above two cases one by one.

Rule 4a: Shifting Summing Point after the Block

Consider the block diagram shown in the following figure. Here, the summing point is present before the block.

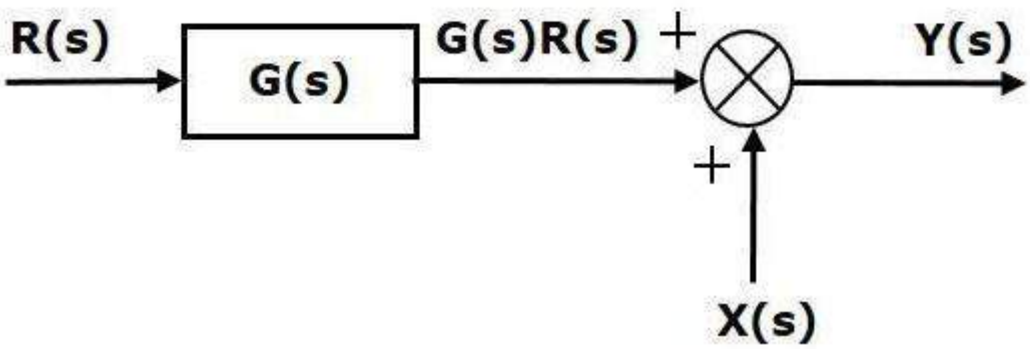


Summing point has two inputs R(s) and X(s). The output of it is {R(s)+X(s)}
So, the input to the block G(s) is {R(s)+X(s)} and the output of it is –

$$Y(s) = G(s)\{R(s)+X(s)\}$$

$$\Rightarrow Y(s) = G(s)R(s) + G(s)X(s) \quad \textbf{(Equation 1)}$$

Now, shift the summing point after the block. This block diagram is shown in the following figure.



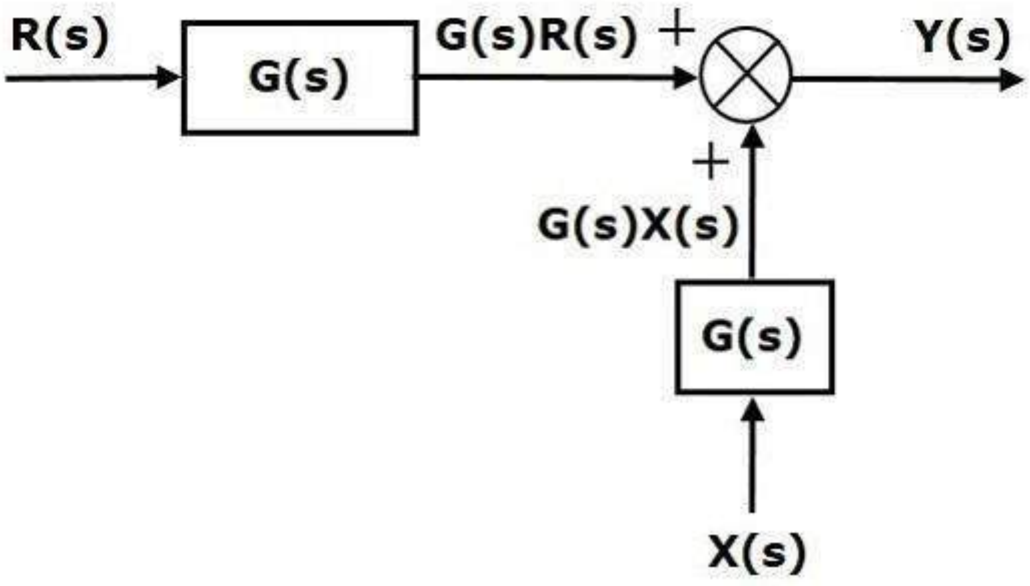
Output of the block G(s) is G(s)R(s)

The output of the summing point is

$Y(s)=G(s)R(s)+X(s)$ **(Equation 2)**

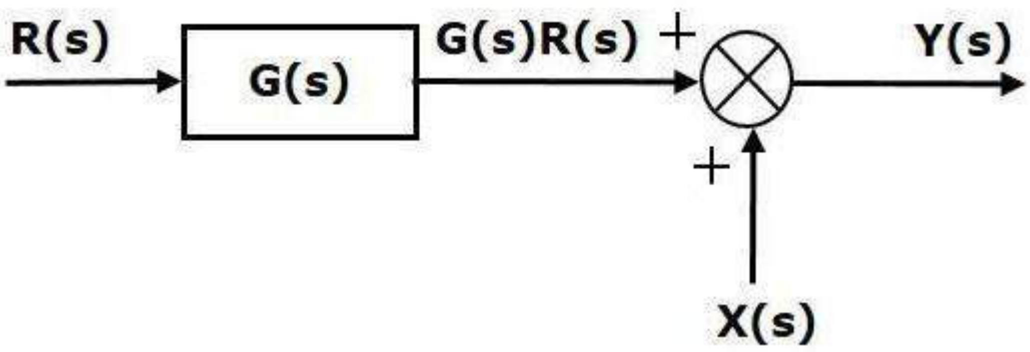
Compare Equation 1 and Equation 2.

The first term 'G(s)R(s)' is same in both the equations. But, there is difference in the second term. In order to get the second term also same, we require one more block G(s). It is having the input X(s) and the output of this block is given as input to summing point instead of X(s). This block diagram is shown in the following figure.



Rule 4b: Shifting Summing Point Before the Block

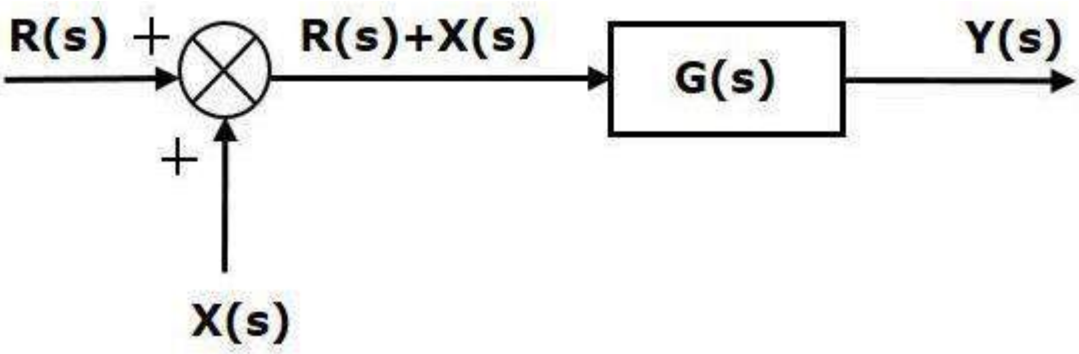
Consider the block diagram shown in the following figure. Here, the summing point is present after the block.



Output of this block diagram is -

$Y(s) = G(s)\,R(s) + X(s)$ **(Equation 3)**

Now, shift the summing point before the block. This block diagram is shown in the following figure.

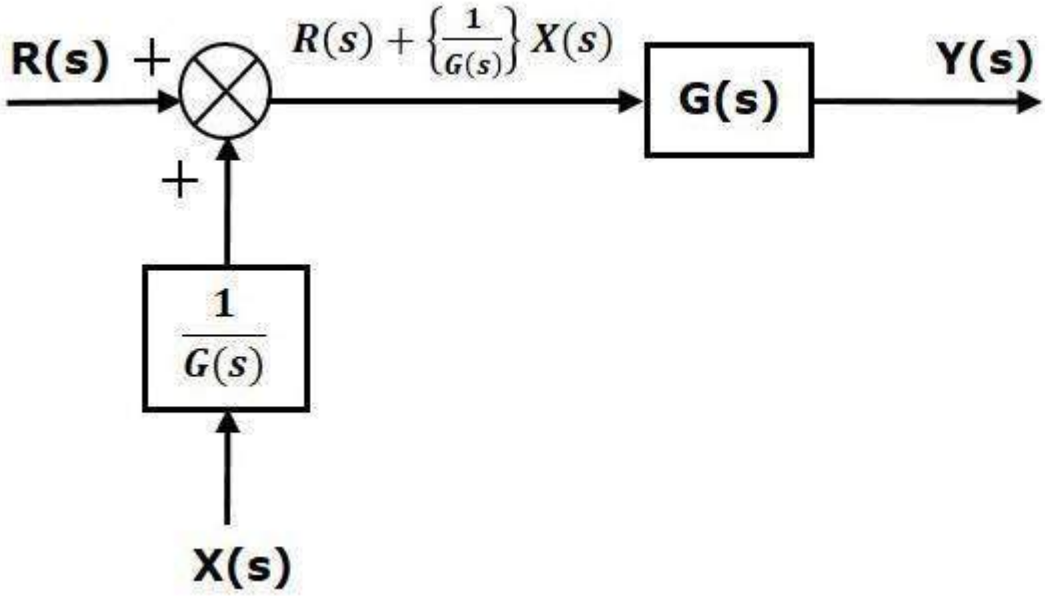


Output of this block diagram is -

$Y(S) = G(s) R(s) + G(s) X(s)$ **(Equation 4)**

Compare Equation 3 and Equation 4,

The first term 'G(s) R(s)′ is same in both equations. But, there is difference in the second term. In order to get the second term also same, we require one more block 1/G(s). It is having the input X(s) and the output of this block is given as input to summing point instead of X(s). This block diagram is shown in the following figure.



Rule 5: Block Diagram Algebra for Take-off Points

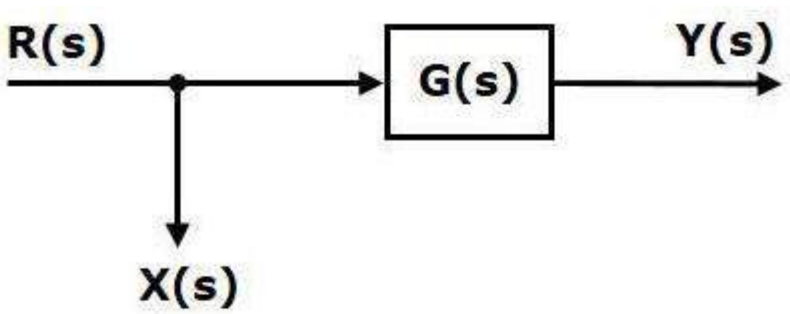
There are two possibilities of shifting the take-off points with respect to blocks −

- Shifting take-off point after the block
- Shifting take-off point before the block

Let us now see what kind of arrangements are to be done in the above two cases, one by one.

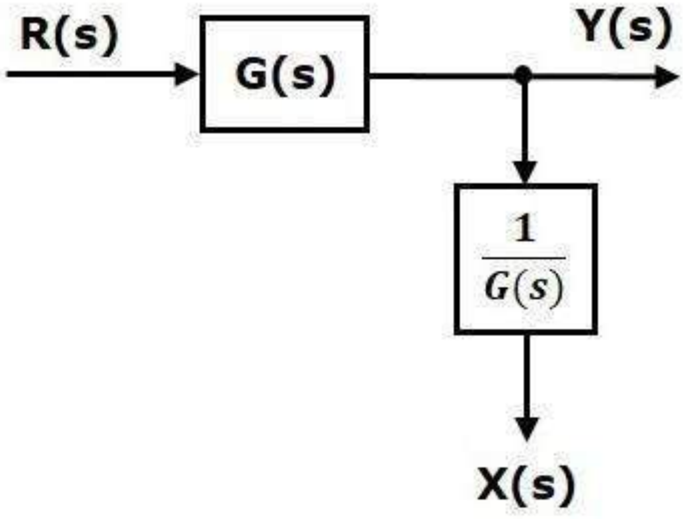
Rule5a: Shifting Take-off Point after the Block

Consider the block diagram shown in the following figure. In this case, the take-off point is present before the block.



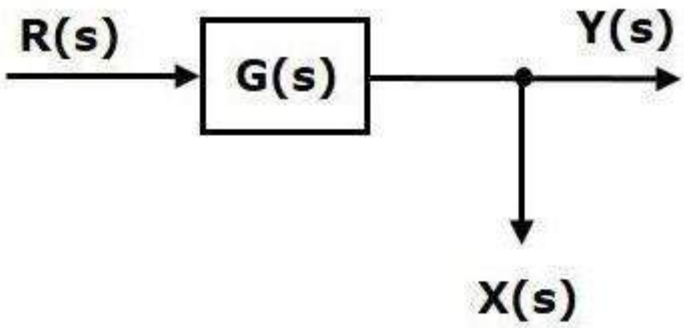
Here, $X(s) = R(s)$ and $Y(s) = G(s)R(s)$

When you shift the take-off point after the block, the output Y(s) will be same. But, there is difference in X(s) value. So, in order to get the same X(s) value, we require one more block 1/G(s). It is having the input Y(s) and the output is X(s). This block diagram is shown in the following figure.



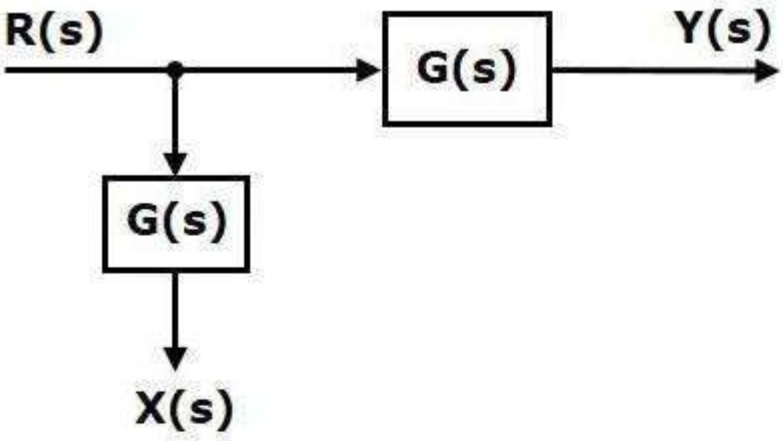
Rule 5b: Shifting Take-off Point Before the Block

Consider the block diagram shown in the following figure. Here, the take-off point is present after the block.



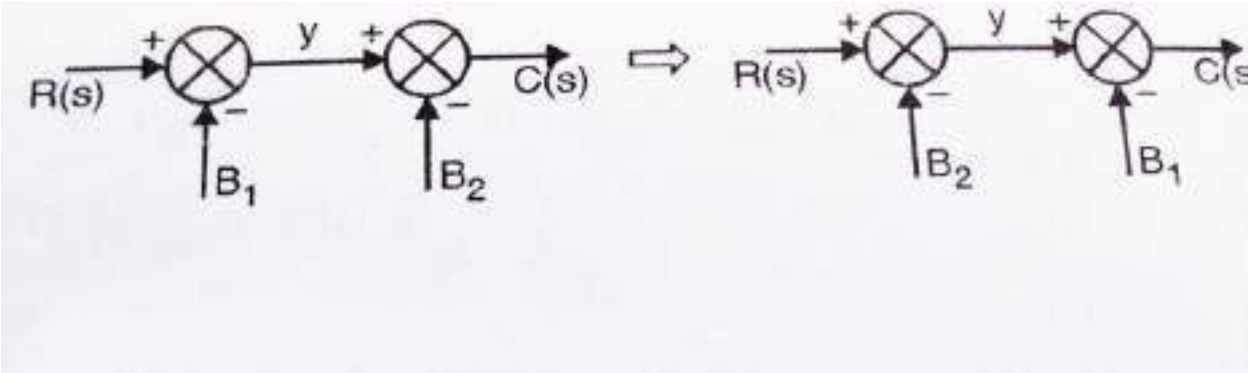
Here, X(s) = Y(s) = G(s)R(s)

When you shift the take-off point before the block, the output Y(s) will be same. But, there is difference in X(s) value. So, in order to get same X(s) value, we require one more block G(s). It is having the input R(s) and the output is X(s). This block diagram is shown in the following figure.



Rule 6: Associative Law For Summing Point

This can be better explained by taking below diagram



Y= R(s) – B1
C(s) = y – B2 = R(s) – B1 – B2
This law is applicable only to summing points which are connected directly to each other.
Note: If there is a block present between two summing points(and hence they are not connected directly) then this rule can't be applied.

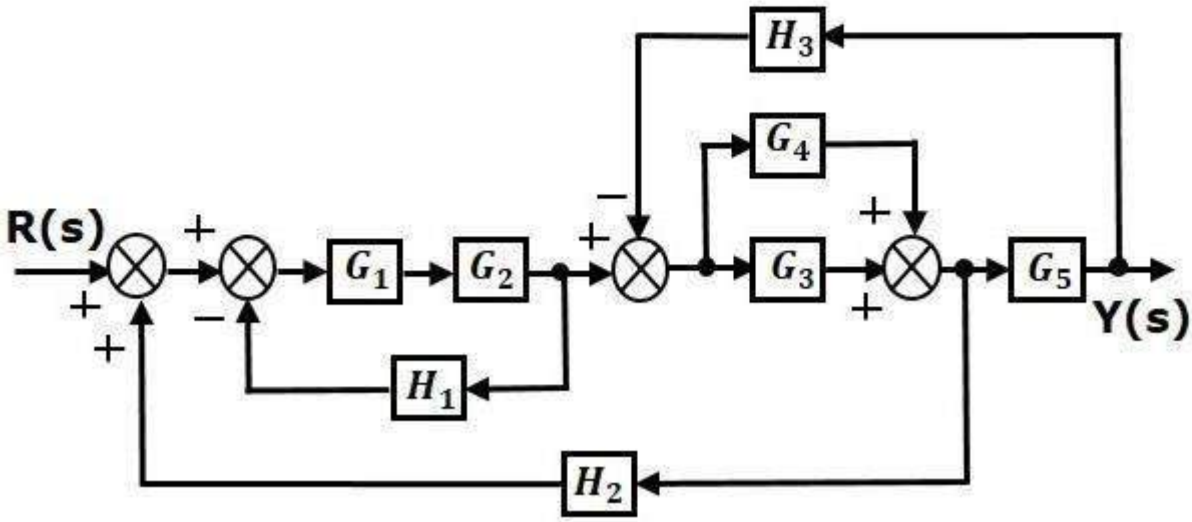
Procedure for finding TF by using Block Diagram Reduction Rules

Follow these rules for simplifying (reducing) the block diagram, which is having many blocks, summing points and take-off points.

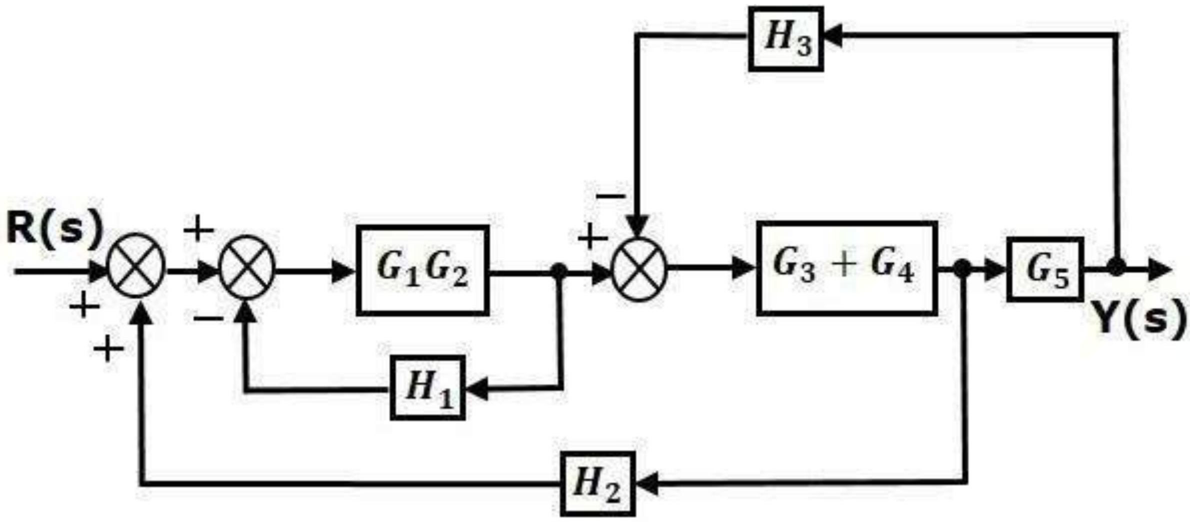
- **Rule 1** − Check for the blocks connected in series and simplify.
- **Rule 2** − Check for the blocks connected in parallel and simplify.
- **Rule 3** − Check for the blocks connected in feedback loop and simplify.
- **Rule 4** − If there is difficulty with take-off point while simplifying, shift it towards right or left of the given block which one is suitable.
- **Rule 5** − If there is difficulty with summing point while simplifying, shift it towards right or left of the given block which one is suitable.
- **Rule 6** − Repeat the above steps till you get the simplified form, i.e., single block.

Example

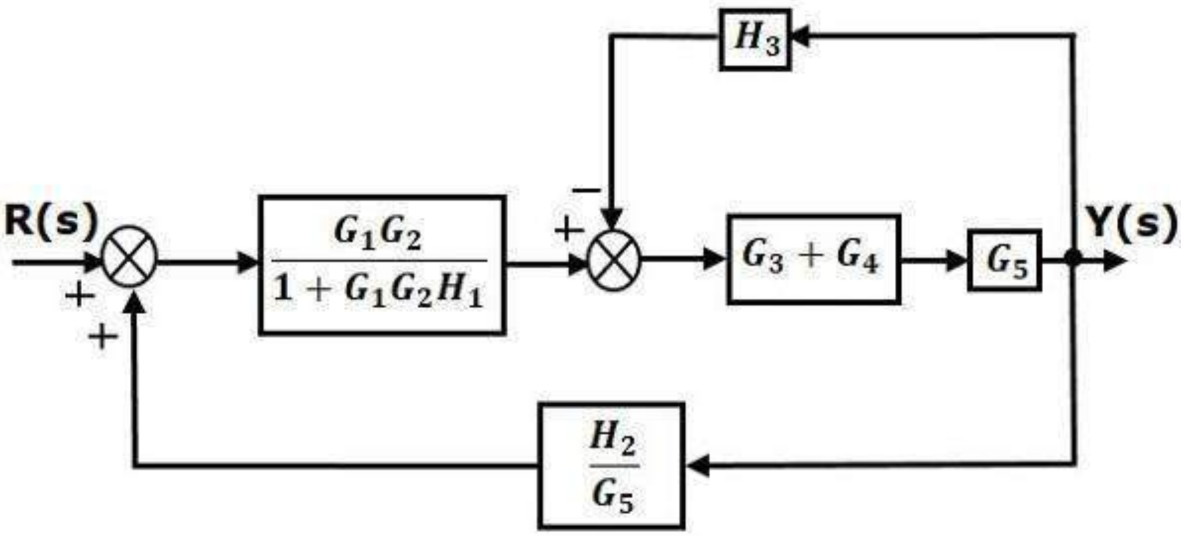
Consider the block diagram shown in the following figure. Let us simplify (reduce) this block diagram using the block diagram reduction rules.



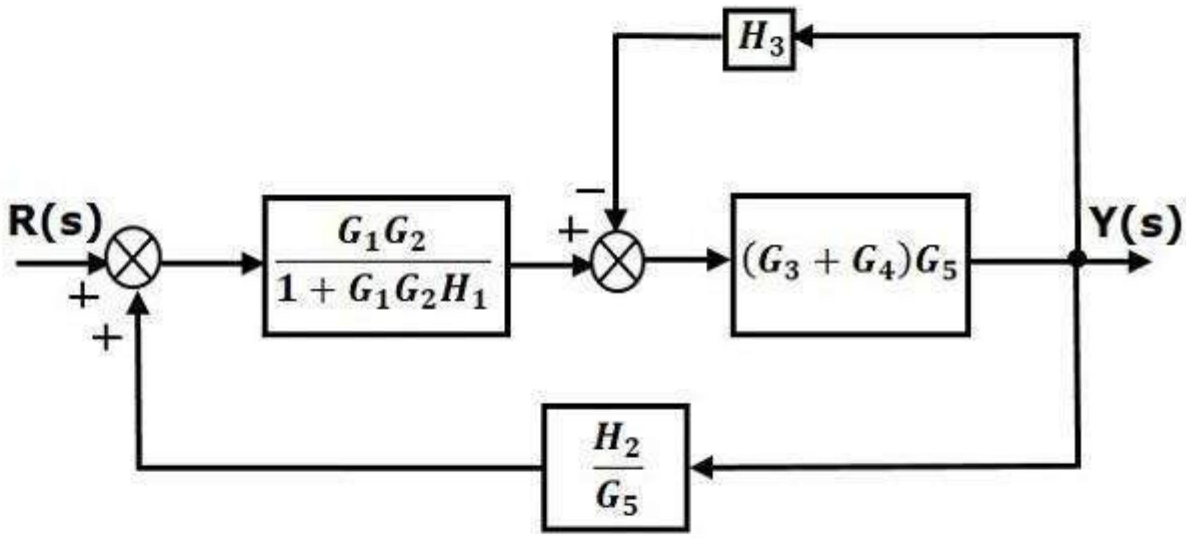
Step 1 − Use Rule 1 for blocks G1 and G2. Use Rule 2 for blocks G3 and G4. The modified block diagram is shown in the following figure.



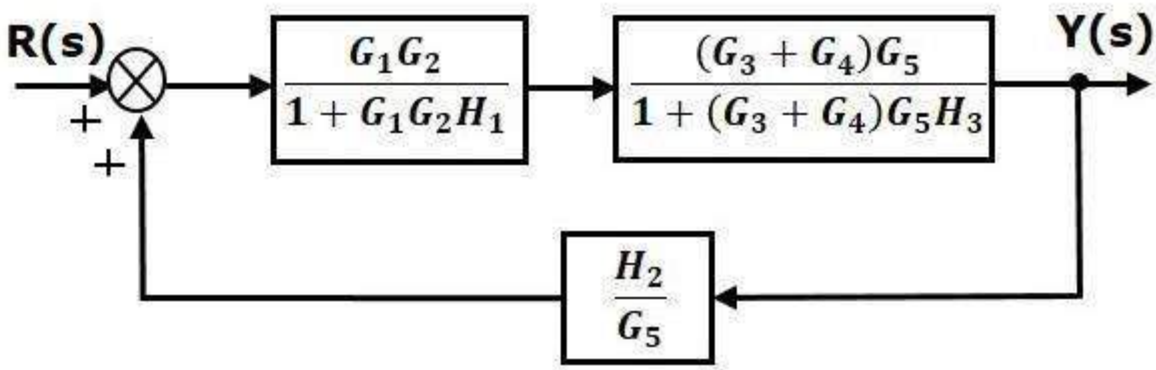
Step 2 − Use Rule 3 for blocks G1G2 and H1. Use Rule 4 for shifting take-off point after the block G5. The modified block diagram is shown in the following figure.



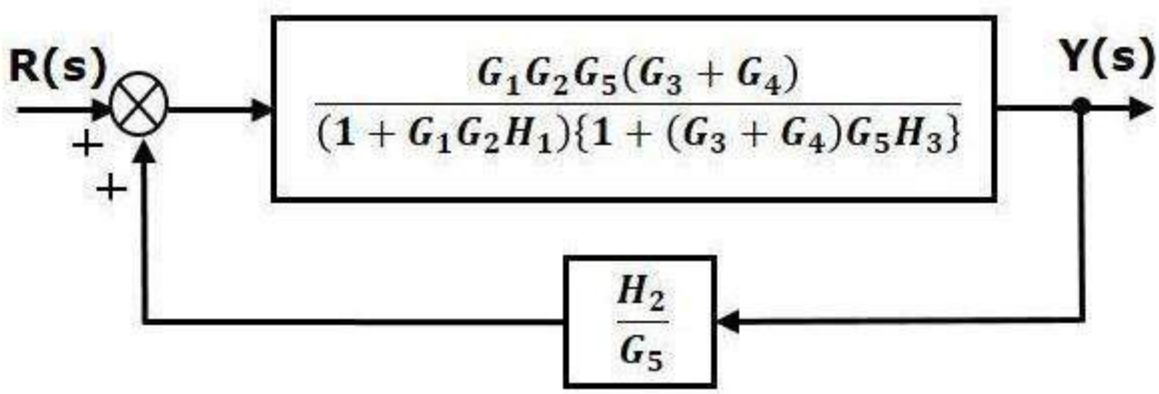
Step 3 − Use Rule 1 for blocks (G3+G4) and G5. The modified block diagram is shown in the following figure.



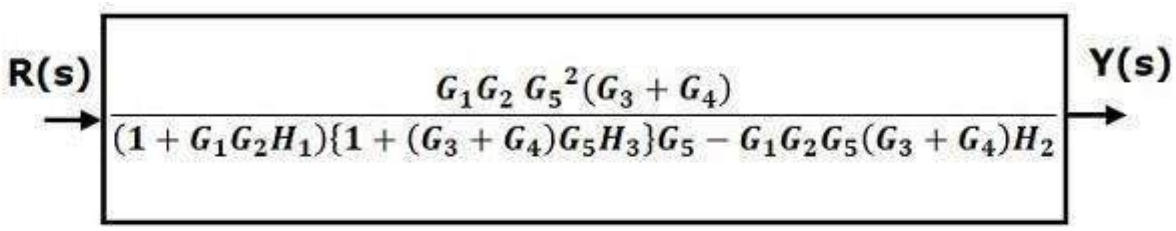
Step 4 − Use Rule 3 for blocks (G3+G4)G5 and H3. The modified block diagram is shown in the following figure.



Step 5 − Use Rule 1 for blocks connected in series. The modified block diagram is shown in the following figure.



Step 6 − Use Rule 3 for blocks connected in feedback loop. The modified block diagram is shown in the following figure. This is the simplified block diagram.



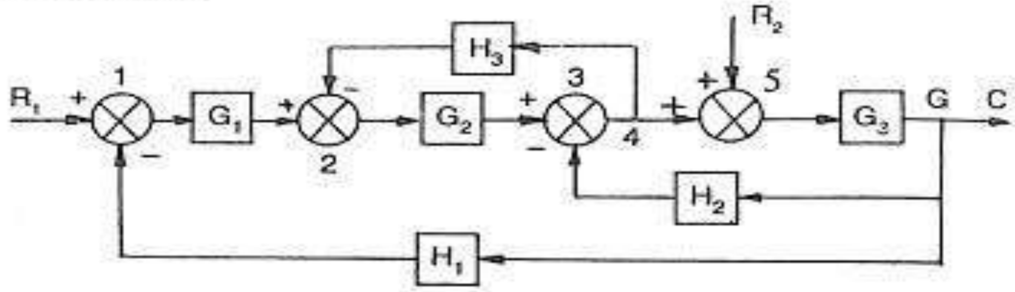
Therefore, the transfer function of the system is

$Y(s)/R(s)=G_1G_2G_5{}^2\,(G_3+G_4)/(1+G_1G_2\,H_1)\{1+(\,G_3+G_4)\,G_5H_3\}G_5-G_1G_2G_5\,(G_3+G_4)H_2$

Note − Follow these steps in order to calculate the transfer function of the block diagram having multiple inputs.

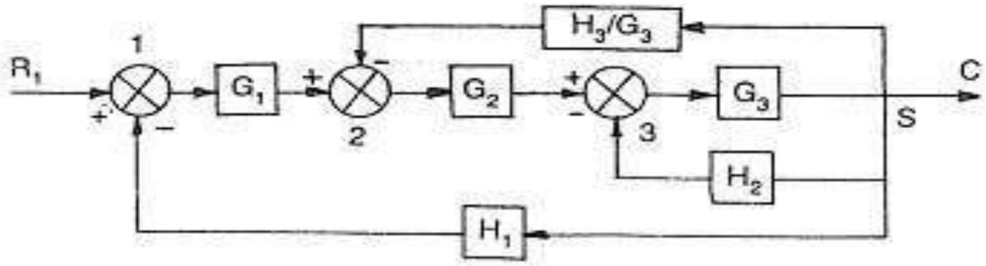
- **Step 1** − Find the transfer function of block diagram by considering one input at a time and make the remaining inputs as zero.
- **Step 2** − Repeat step 1 for remaining inputs.
- **Step 3** − Get the overall transfer function by adding all those transfer functions.

Problem Evaluate $\frac{C}{R_1}$ and $\frac{C}{R_2}$ for a system whose block diagram representation is shown in Fig. R_1 is the input to summing point No. 1.

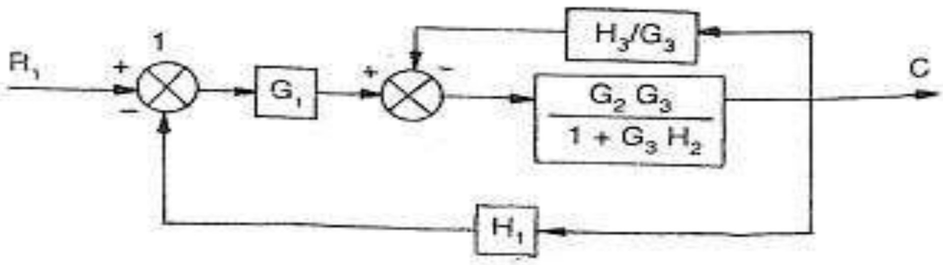


Solution

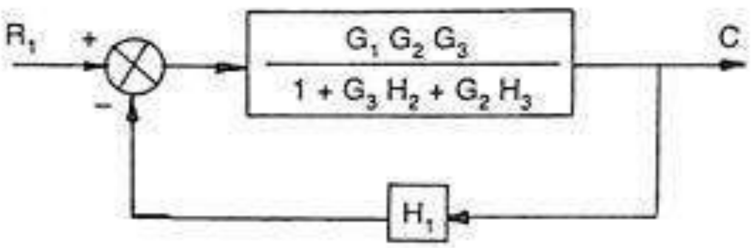
Evaluation of C/R₁ Assume $R_2 = 0$. Therefore summing point No. 5 can be removed. Shift take off point No. 4 beyond block G_3



Eliminate the feedback loop between points **3 and 6**



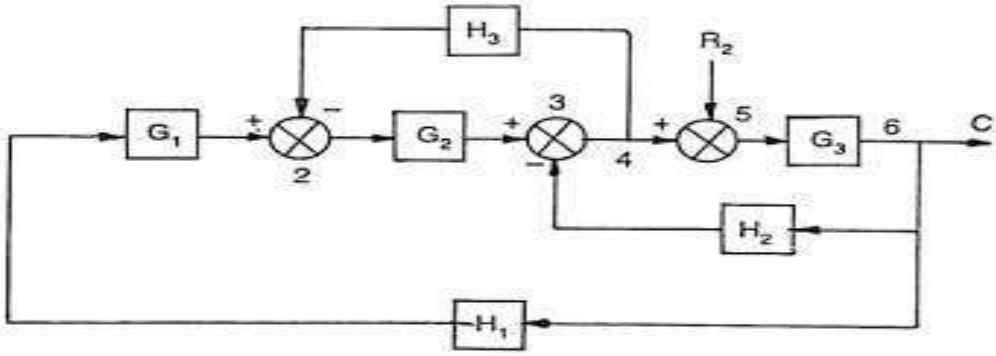
Eliminating the feed back loop again



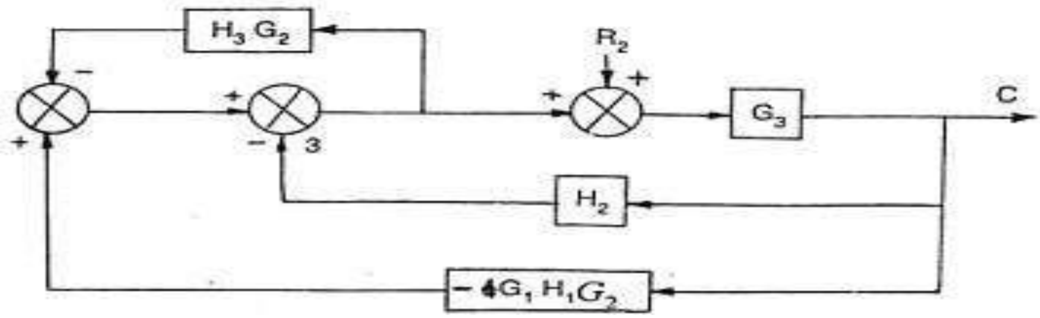
$$\frac{C}{R_1} = \frac{G_1G_2G_3}{1 + G_3H_2 + H_3G_2 + G_1G_2G_3H_1} \quad \textbf{Ans.}$$

Evaluation of C/R_2

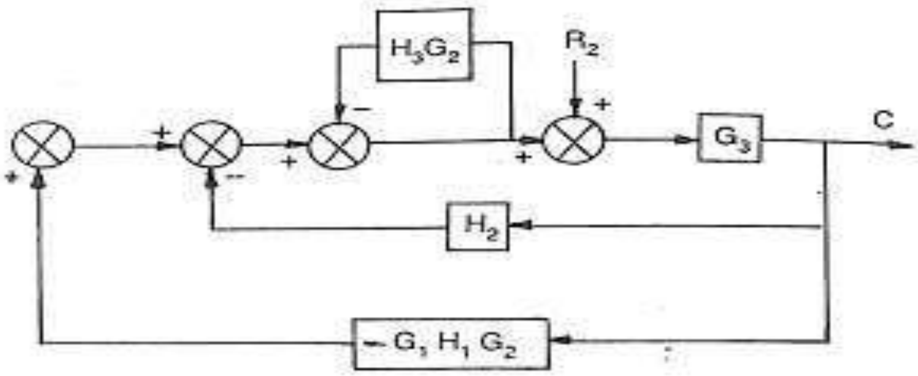
Assume $R_1 = 0$. Thus summing point No. 1 can be removed



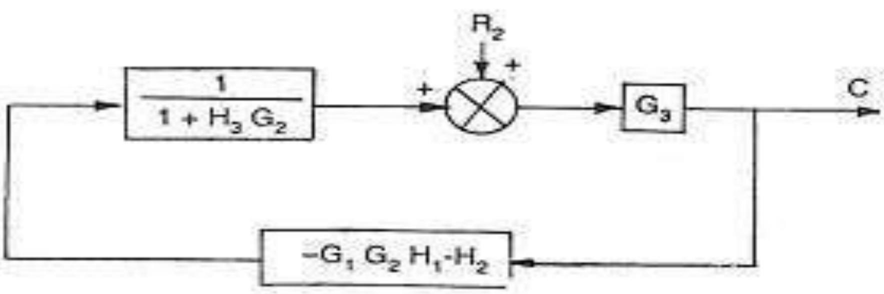
Shifting the summing point No. 2 and rearranging beyond G_2



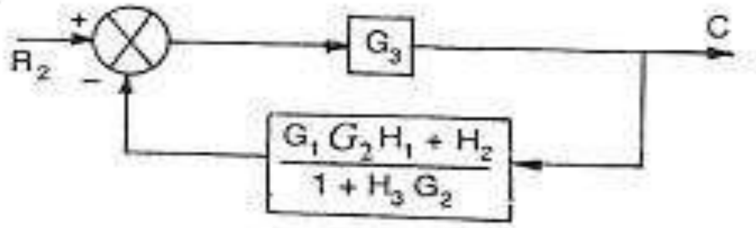
Rearranging, we get



Rearranging and eliminating the feedback loop



Rearranging,



Eliminating the feedback loop, we get

$$\frac{C}{R_2} = \frac{G_3(1 + H_3G_2)}{1 + H_3G_2 + G_3(G_1G_2H_1 + H_2)} \qquad \text{Ans.}$$

The block diagram reduction process takes more time for complicated systems. Because, we have to draw the (partially simplified) block diagram after each step. So, to overcome this drawback, use signal flow graphs (representation).

SIGNAL FLOW GRAPHS

Signal flow graph is a graphical representation of algebraic equations. In this chapter, let us discuss the basic concepts related signal flow graph and also learn how to draw signal flow graphs.

Characteristics of SFG: SFG is a graphical representation of the relationship between the variables of a set of linear algebraic equations. It doesn't require any reduction technique or process.

- It represents a network in which nodes are used for the representation of system variable which is connected by direct branches.
- SFG is a diagram which represents a set of equations. It consists of nodes and branches such that each branch of SFG having an arrow which represents the flow of the signal.
- It is only applicable to the linear system.

Terminology used in SFG

Nodes and branches are the basic elements of signal flow graph.

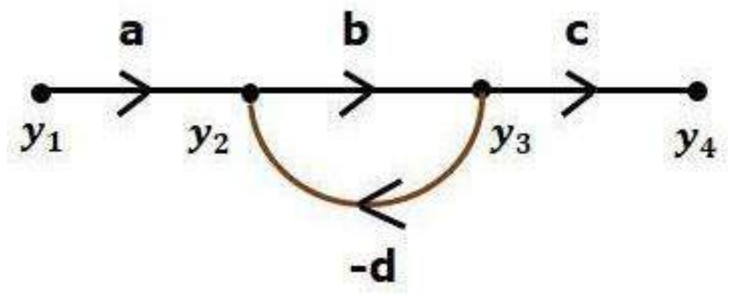
1. Node

Node is a point which represents either a variable or a signal. There are three types of nodes — input node, output node and mixed node.

- **Input Node or source**− It is a node, which has only outgoing branches.
- **Output Node or sink** − It is a node, which has only incoming branches.
- **Mixed Node** − It is a node, which has both incoming and outgoing branches.

Example

Let us consider the following signal flow graph to identify these nodes.



- The **nodes** present in this signal flow graph are $\mathbf{y_1}$**,** $\mathbf{y_2}$**,** $\mathbf{y_3}$ and $\mathbf{y_4}$.
- $\mathbf{y_1}$ and $\mathbf{y_4}$ are the **input node** and **output node** respectively.
- $\mathbf{y_2}$ and $\mathbf{y_3}$ are **mixed nodes**.

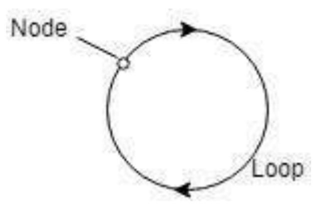
2. Branch

Branch is a line segment which joins two nodes. It has both **gain** and **direction**. For example, there are four branches in the above signal flow graph. These branches have **gains** of **a, b, c** and **-d**.

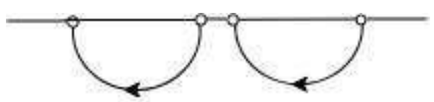
3. Forward Path

It is a path from an input node to an output node in the direction of branch arrow.

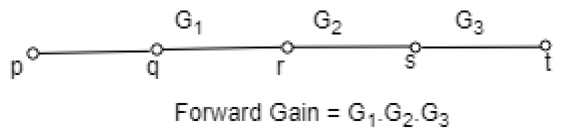
4. **Loop**: It is a path that starts and ends at the same node.



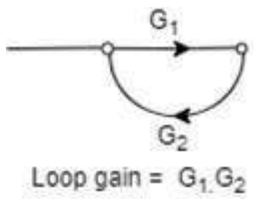
5. **Non-touching loop**: Loop is said to be non-touching if they do not have any common node.



6. **Forward path gain**: A product of all branches gain along the forward path is called Forward path gain.



7. **Loop Gain**: Loop gain is the product of branch gain which travels in the loop.



Construction of Signal Flow Graph

Let us construct a signal flow graph by considering the following algebraic equations −

$$y_2 = a_{12}\, y_1 + a_{42}\, y_4$$

$$y_3 = a_{23}\, y_2 + a_{53}\, y_5$$

$$y_4 = a_{34}\, y_3$$

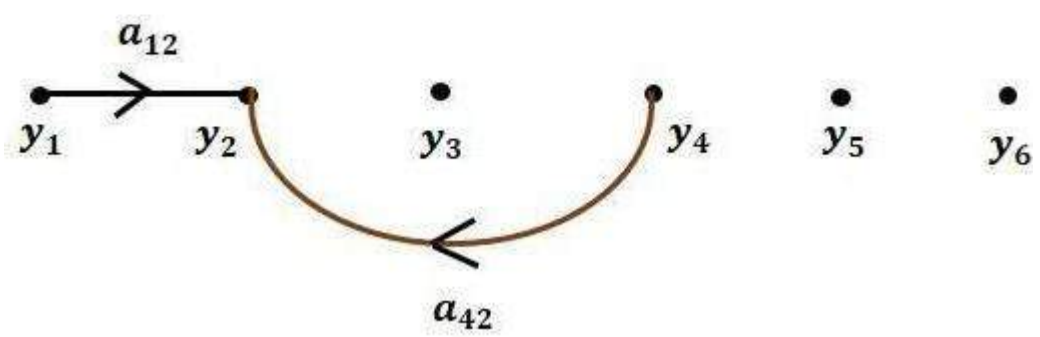
$$y_5 = a_{45}\, y_4 + a_{35}\, y_3$$

$$y_6 = a_{56}\, y_5$$

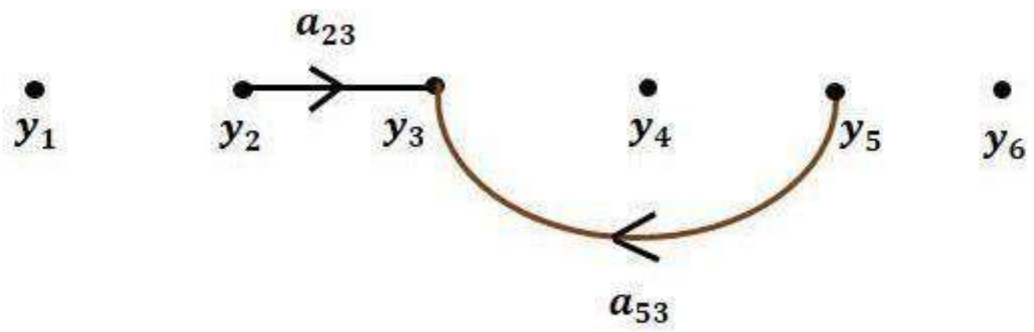
There will be six **nodes** (y_1, y_2, y_3, y_4, y_5 and y_6) and eight **branches** in this signal flow graph. The gains of the branches are a_{12}, a_{23}, a_{34}, a_{45}, a_{56}, a_{42}, a_{53} and a_{35}.

To get the overall signal flow graph, draw the signal flow graph for each equation, then combine all these signal flow graphs and then follow the steps given below −

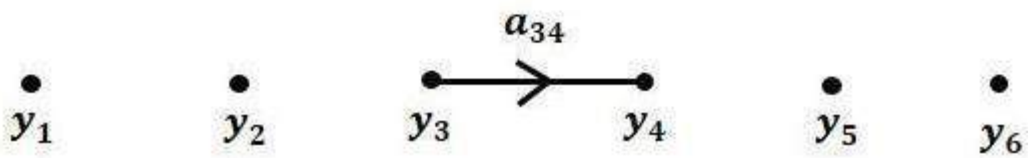
Step 1 – Signal flow graph for $y_2 = a_{12}\, y_1 + a_{42}\, y_4$ is shown in the following figure.



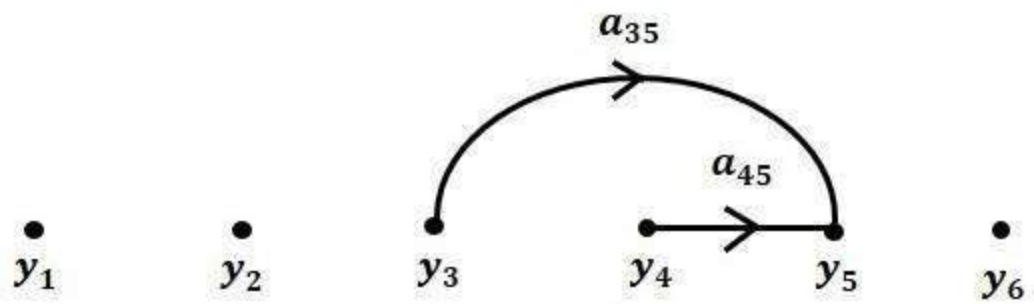
Step 2 – Signal flow graph for $y_3 = a_{23}\, y_2 + a_{53}\, y_5$ is shown in the following figure.



Step 3 − Signal flow graph for $y_4 = a_{34}\, y_3$ is shown in the following figure.

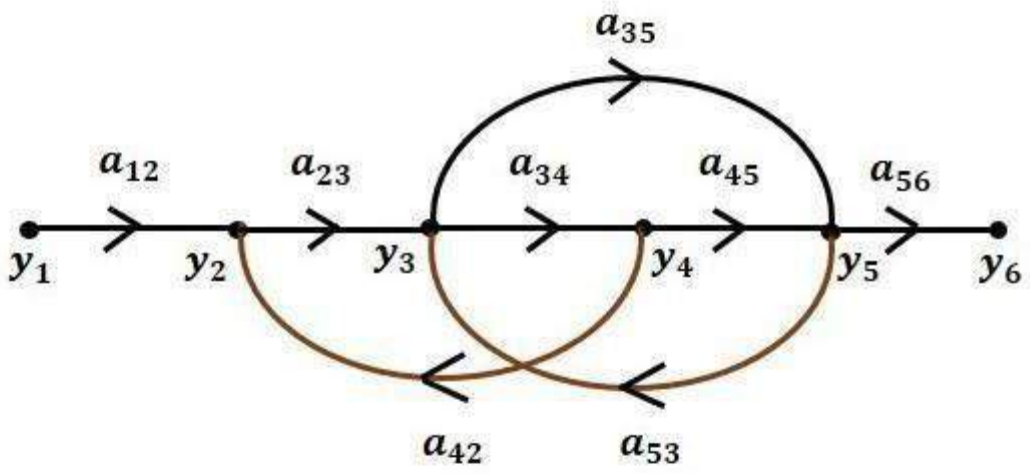


Step 4 − Signal flow graph for $y_5 = a_{45}\, y_4 + a_{35}\, y_3$ is shown in the following figure.



Step 5 − Signal flow graph for $y_6 = a_{56}\, y_5$ is shown in the following figure.

Step 6 − Signal flow graph of overall system is shown in the following figure.



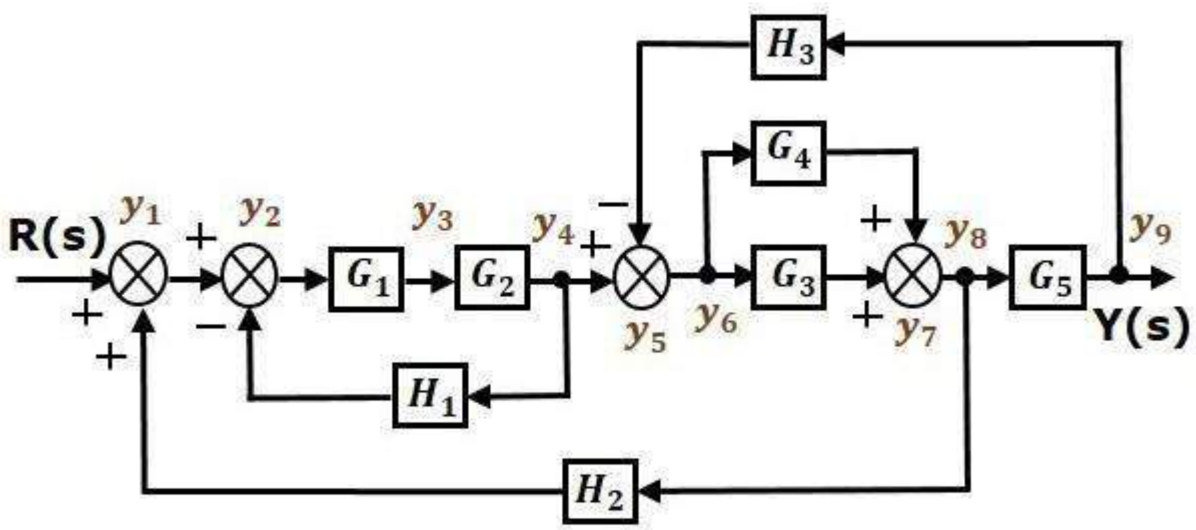
Conversion of Block Diagrams into Signal Flow Graphs

Follow these steps for converting a block diagram into its equivalent signal flow graph.

- Represent all the signals, variables, summing points and take-off points of block diagram as **nodes** in signal flow graph.
- Represent the blocks of block diagram as **branches** in signal flow graph.
- Represent the transfer functions inside the blocks of block diagram as **gains** of the branches in signal flow graph.
- Connect the nodes as per the block diagram. If there is connection between two nodes (but there is no block in between), then represent the gain of the branch as one. **For example**, between summing points, between summing point and takeoff point, between input and summing point, between take-off point and output.
-

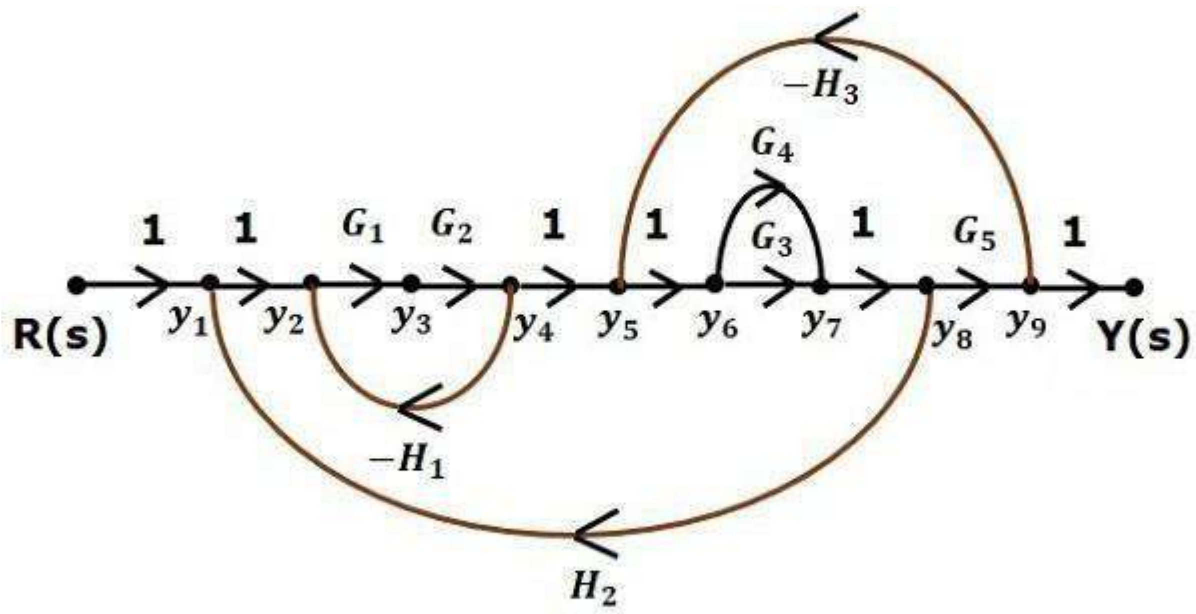
Example

Let us convert the following block diagram into its equivalent signal flow graph.



Represent the input signal R(s) and output signal C(s) of block diagram as input node R(s) and output node C(s) of signal flow graph.
Just for reference, the remaining nodes (y_1 to y_9) are labelled in the block diagram. There are nine nodes other than input and output nodes. That is four nodes for four summing points, four nodes for four take-off points and one node for the variable between blocks G1 and G2.

The following figure shows the equivalent signal flow graph.

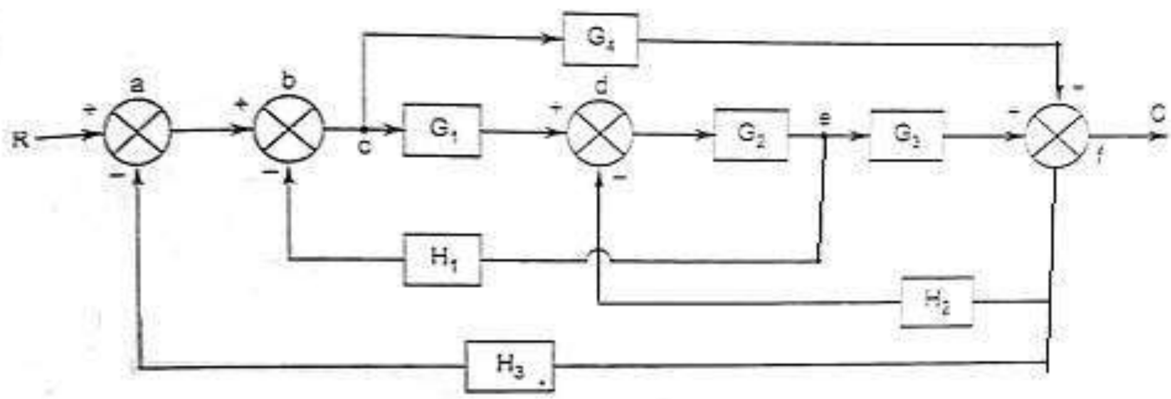


With the help of Mason's gain formula (discussed in the next chapter), you can calculate the transfer function of this signal flow graph. This is the advantage of signal flow graphs. Here, we no need to simplify (reduce) the signal flow graphs for calculating the transfer function.

Note: 1. If summing point is present before a take off point it may be assume as same node.

2. If there is a present of summing point in series (no block with in it), it may be take ias same node.

Ex: Determine transfer function by using Mason's gain formula.



Solution:

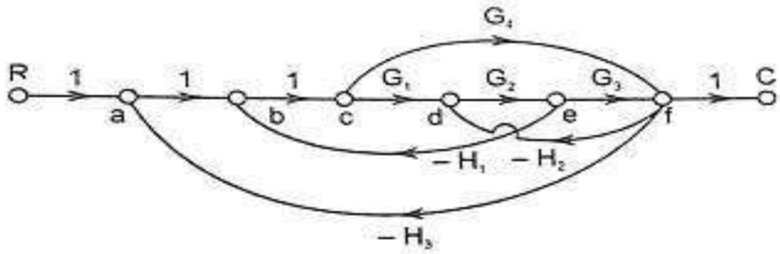


Fig. E3.13(a) *SFG*

Step I : Forward paths of the SFG are as follows :

(*i*) $a - b - c - d - e - f$ $\quad\therefore\quad P_1 = G_1\,G_2\,G_3$

(*ii*) $a - b - c - f$ $\quad\therefore\quad P_2 = G_4$

Step II : Individual loops of the SFG are as follows :

(*i*) $d - e - f - d$ $\quad\therefore\quad L_1 = -\,G_2\,G_3\,H_2$

(*ii*) $b - c - d - e - b$ $\quad\therefore\quad L_2 = -\,G_1\,G_2\,H_1$

(*iii*) $a - b - c - f - a$ $\quad\therefore\quad L_3 = -\,G_4\,H_3$

(*iv*) $b - c - f - d - e - b$ $\quad\therefore\quad L_4 = G_4\,H_2\,G_2\,H_1$

(*v*) $a - b - c - d - e - f - a$ $\quad\therefore\quad L_5 = -\,G_1\,G_2\,G_3\,H_3$

Step III : Gain products of all possible two non-touching loops are as follows :

Loop L_1 and Loop L_2 are non-touching loops

$$\therefore \quad L_{12} = G_2 H_1 G_4 H_2$$

Step IV : $\quad \Delta = 1 - (L_1 + L_2 + L_3 + L_4) + L_{12}$

Step V : For $\quad P_1, \Delta_1 = 1$

$$\therefore \frac{C(s)}{R(s)} = \frac{G_1\,G_2\,G_3\,G_4\,G_5}{1 - \left(\begin{matrix} G_2\,H_1 + G_4\,H_2 + G_2\,G_3\,G_4\,G_5\,G_6\,G_7\,G_8 \\ + G_2\,G_3\,G_4\,G_5\,G_6\,G_8\,H_4 \end{matrix}\right) + G_1\,G_4\,H_1\,H_2} \text{. **Ans.**}$$

Mason's Gain Formula

Let us now discuss the Mason's Gain Formula. Suppose there are 'N' forward paths in a signal flow graph. The gain between the input and the output nodes of a signal flow graph is nothing but the **transfer function** of the system. It can be calculated by using Mason's gain formula.

Mason's gain formula is

$$T = C(s)/R(s) = \left(\frac{1}{\Delta}\right) \sum_{i=1}^{N} P_i \Delta_i$$

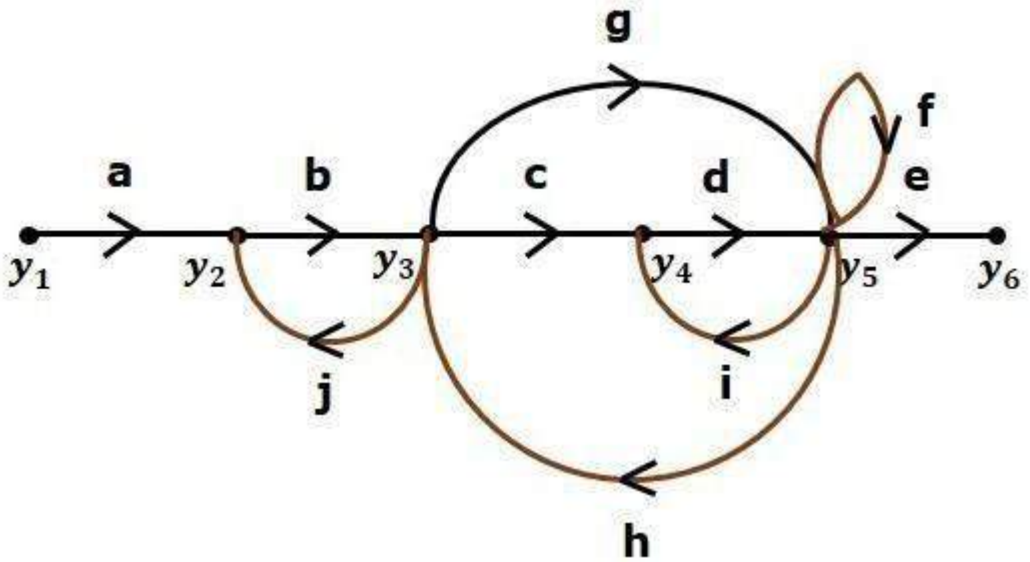
Where,

- **C(s)** is the output node
- **R(s)** is the input node
- **T** is the transfer function or gain between R(s)R(s) and C(s)C(s)
- $\mathbf{P_i}$ is the i^{th} forward path gain

Δ=1−(sum of all individual loop gains) + (sum of gain products of all possible two non touching loops) − (sum of gain products of all possible three non touching loops)+...

Δ_i is obtained from Δ by removing the loops which are touching the i^{th} forward path.

Consider the following signal flow graph in order to understand the basic terminology involved here.



Path

It is a traversal of branches from one node to any other node in the direction of branch arrows. It should not traverse any node more than once.

Examples y2→y3→y4→y5 and y5→y3→y2

Forward Path

The path that exists from the input node to the output node is known as **forward path**.

Examples − y1→y2→y3→y4→y5→y6 and y1→y2→y3→y5→y6.

Forward Path Gain

It is obtained by calculating the product of all branch gains of the forward path.

Examples – abcde is the forward path gain of y1→y2→y3→y4→y5→y6 and abge is the forward path gain of y1→y2→y3→y5→y6.

Loop

The path that starts from one node and ends at the same node is known as **loop**. Hence, it is a closed path.

Examples y2→y3→y2 and y3→y5→y3.

Loop Gain

It is obtained by calculating the product of all branch gains of a loop.

Examples − bj is the loop gain of y2→y3→y2 and gh is the loop gain of y3→y5→y3.

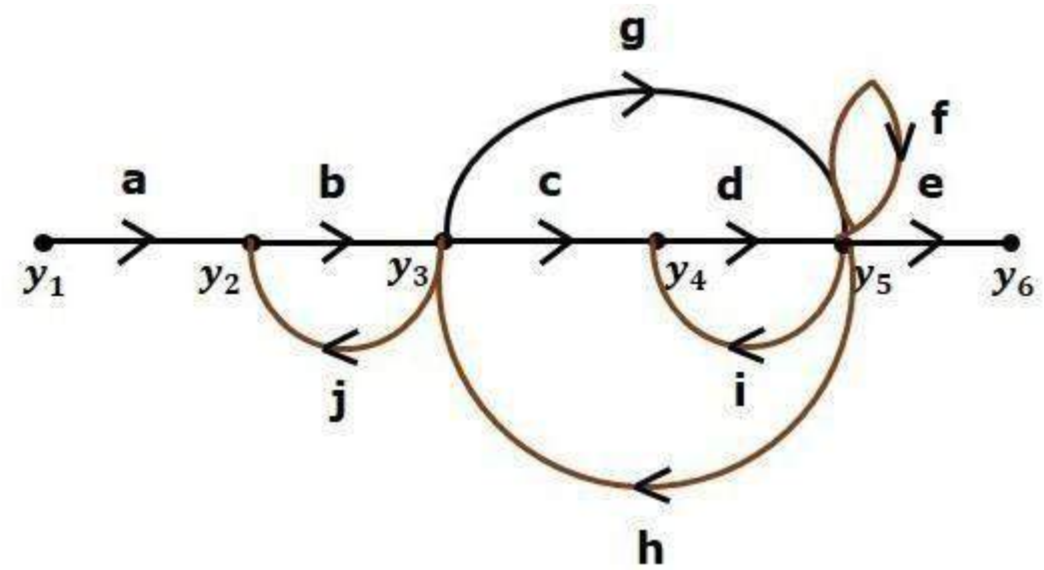
Non-touching Loops

These are the loops, which should not have any common node.

Examples − The loops $y2 \rightarrow y3 \rightarrow y2$ and $y4 \rightarrow y5 \rightarrow y4$ are non-touching.

Calculation of Transfer Function using Mason's Gain Formula

Let us consider the same signal flow graph for finding transfer function.



Number of forward paths, N = 2.

- First forward path is - $y1 \rightarrow y2 \rightarrow y3 \rightarrow y4 \rightarrow y5 \rightarrow y6$.
- First forward path gain, $p1=abcde$.
- Second forward path is - $y1 \rightarrow y2 \rightarrow y3 \rightarrow y5 \rightarrow y6$.
- Second forward path gain, $p2=abge$.

Number of individual loops, L = 5.

Loops are - $y2 \rightarrow y3 \rightarrow y2$, $y3 \rightarrow y5 \rightarrow y3$, $y3 \rightarrow y4 \rightarrow y5 \rightarrow y3$, $y4 \rightarrow y5 \rightarrow y4$ and $y5 \rightarrow y5$.
Loop gains are - $l1=bj$, $l2=gh$, $l3=cdh$, $l4=di$ and $l5=f$

Number of two non-touching loops = 2.

- First non-touching loops pair is - $y2 \rightarrow y3 \rightarrow y2$, $y4 \rightarrow y5 \rightarrow y4$.
- Gain product of first non-touching loops pair, $l1l4= bjdi$
- Second non-touching loops pair is - $y2 \rightarrow y3 \rightarrow y2$, $y5 \rightarrow y5$
- Gain product of second non-touching loops pair is - $l1l5=bjf$

Higher number of (more than two) non-touching loops are not present in this signal flow graph.

We know,

$\Delta=1-$ (sum of all individual loop gains)
+ (sum of gain products of all possible two non touching loops)
−(sum of gain products of all possible three non touching loops)+...

Substitute the values in the above equation,

$\Delta=1-(bj+gh+cdh+di+f)+(bjdi+bjf)-(0)$
$\Rightarrow\Delta=1-(bj+gh+cdh+di+f)+bjdi+bjf$

There is no loop which is non-touching to the first forward path.

So, $\Delta 1=1$

Similarly, $\Delta 2=1$. Since, no loop which is non-touching to the second forward path.

Substitute, N = 2 in Mason's gain formula

$$T=C(s)R(s)=[P1\Delta 1+P2\Delta 2]/\Delta$$

Substitute all the necessary values in the above equation.

$$T=C(s)R(s)=(abcde)1+(abge)1/[1-(bj+gh+cdh+di+f)+bjdi+bjf]$$

$$\Rightarrow T=C(s)R(s)=(abcde)+(abge)/[1-(bj+gh+cdh+di+f)+bjdi+bjf]$$

Therefore, the transfer function is -

$$T=C(s)R(s)=(abcde)+(abge)/[1-(bj+gh+cdh+di+f)+bjdi+bjf]$$

Course Content

TIME RESPONSE ANALYSIS.

2 . 1Time response of control system.

2 . 2 Standard Test signal: Step signal, Ramp Signal, Parabolic signal, Impulse Signal

2 . 3 Time Response of first order system with: Unit step response, Unit impulse response.

2 . 4 Time response of second order system to the unit step input.

2.5 Time response specification.

Derivation of expression for rise time, peak time, peak overshoot, settling

time and steady state error.

2.6 Steady state error and error constants.

2. 7 Types of control system.[Steady state errors in Type-0, Type-1, Type-2 system

2. 8 Effect of adding poles and zero to transfer function.

CHAPTER II

Time Response Analysis

The variation of output with respect to time is known as time response. The time response consists of two parts.

- Transient response
- Steady state response

Here, both the transient and the steady states are indicated in the figure 1. The responses corresponding to these states are known as transient and steady state responses.

Mathematically, we can write the time response c(t) as

$$C(t)=Ctr(t)+Css(t) \quad (1)$$

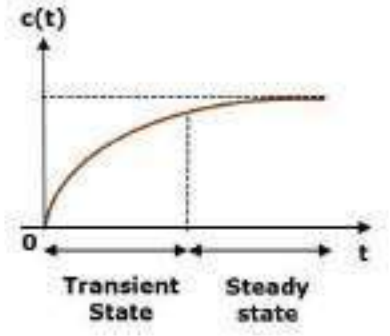


Figure 1 Time response of a system

Where,

- $c_{tr}(t)$ is the transient response
- $c_{ss}(t)$ is the steady state response

Transient Response

The transient response is the part of the tie response which goes to zero after large interval of time 't'. Ideally, this value of 't' is infinity and practically, it is five times constant.

Mathematically, we can write it as

$$\lim_{t\to\infty} Ctr(t)=0$$

Steady state Response

The part of the time response that remains even after the transient response has zero value for large values of 't' is known as **steady state response**. This means, the transient response will be zero even during the steady state.

Example

Let us find the transient and steady state terms of the time response of the control system c(t)=10+5e−t

Here, the second term 5e−t will be zero as **t** denotes infinity. So, this is the **transient term**. And the first term 10 remains even as **t** approaches infinity. So, this is the **steady state term**.

Standard Test Signals

The standard test signals are impulse, step, ramp and parabolic. These signals are used to know the performance of the control systems using time response of the output.

Unit Impulse Signal

A signal which has zero value everywhere except at t= 0, where its magnitude is infinite. It is also known as δ-function. Mathematically:

$\delta(t) = 0 \;\; ; t\neq 0$

$\quad = \infty \; ; t = 0 \qquad (2)$

and $\int_{-\epsilon}^{+\epsilon} \delta(t)dt = 1$ where ϵ tends to zero

The figure (2a) shows unit impulse signal.

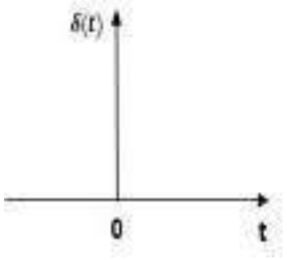


Figure (2a)

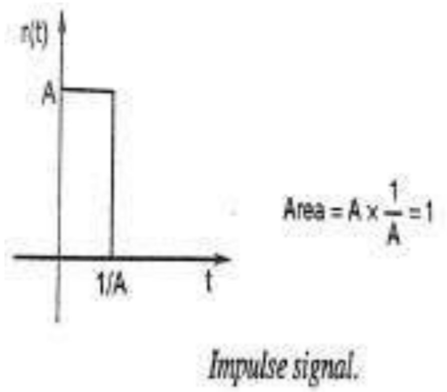


Figure (2b)

Practically a perfect impulse signal cannot be achieved. It is generally approximated be a pulse of unit area as shown in figure (2b).

An unit impulse signal is the derivative of a step signal i.e.,

δ (t) = $\frac{d\,u(t)}{dt}$ (3)

Laplace transform of a unit impulse is

L [δ (t)] = L [$\frac{d\,(t)}{dt}$] = s R(s) = 1 (As for step input R(s) = 1/s) (4)

Unit Step Signal

A unit step signal is defined as r(t) = A u(t) (5)

Where u(t)=1; t≥0

0; t<0

u(t) is called as unit step signal.

By taking Laplace transform of r(t), we have

R(s) = A/s (6)

Following figure 3 shows unit step signal.

So, the unit step signal exists for all positive values of 't' including zero

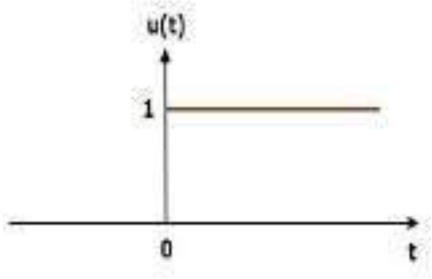


Figure 3: unit step signal

Unit Ramp Signal

A unit ramp signal, r(t) is defined as

$$r(t)= At \; ; t\geq 0$$

$$=0 \; ; t<0 \qquad (7)$$

The ramp signal starts from zero and increases linearly with time. A ramp signal is the integral of a step signal. i.e

Ramp Signal = ∫ step signal

The figure 4 shows unit ramp signal.

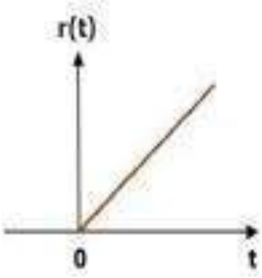


Figure 4 : ramp signal

Unit Parabolic Signal

A unit parabolic signal, r(t) is defined as,

$$r(t)= At^2/2; t\geq 0$$

$$0 \; ; t<0 \qquad (8)$$

By taking the Laplace transform of equation 8,

$$R(s) = A/s^3 \qquad (9)$$

Parabolic signal is integral of a ramp signal. i.e
Parabolic signal = ∫ ramp signal

The figure 5 shows the unit parabolic signal.

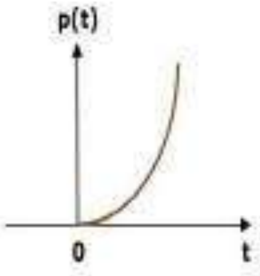


Figure 5: unit parabolic signal

So, the unit parabolic signal exists for all the positive values of **'t'** including zero. And its value increases non-linearly with respect to 't' during this interval. The value of the unit parabolic signal is zero for all the negative values of 't'.

Time Response of the First Order System

Let us discuss the time response of the first order system. Consider the following block diagram of the closed loop control system. Here, an open loop transfer function, 1/sT is connected with a unity negative feedback.

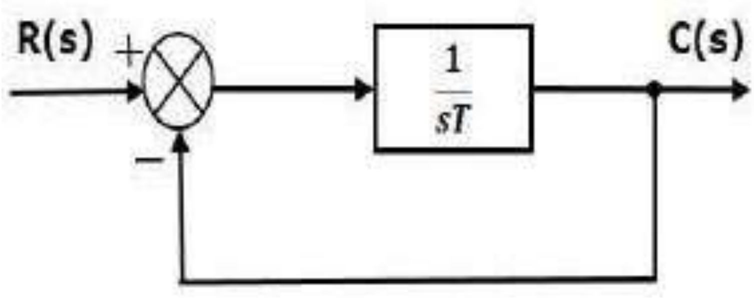


Figure (6) : Block diagram of a first order system

We know that the transfer function of the closed loop control system has unity negative feedback as,

$$\frac{C(s)}{R(s)} = \frac{G(s)}{1+G(s)H(s)}$$

Substitute, G(s)= $\frac{1}{Ts}$ in the above equation.

$$\frac{C(s)}{R(s)} = \frac{\frac{1}{Ts}}{1+\frac{1}{Ts}} = \frac{1}{1+Ts} \qquad (10)$$

The power of s is one in the denominator term. Hence, the above transfer function is of the first order and the system is said to be the **first order system**.

We can re-write the above equation as

$$C(s) = \frac{1}{1+Ts}R(s) \qquad (11)$$

Where,

- **C(s)** is the Laplace transform of the output signal c(t),
- **R(s)** is the Laplace transform of the input signal r(t), and
- **T** is the time constant.

Impulse Response of First Order System

For unit impulse signal R(s) = 1

Consider the equation (11), C(s) = $\frac{1}{1+Ts}$R(s)

Substitute, R(s) = 1 in the above equation.

$$C(s) = \frac{1}{1+Ts} \qquad (12)$$

Rearrange the above equation in one of the standard forms of Laplace transforms.

$$C(s) = \frac{\frac{1}{T}}{s+\frac{1}{T}} \qquad (13)$$

Apply inverse Laplace transform on both sides.

$$c(t)=\frac{1}{T}e^{-\frac{t}{T}} \qquad (14)$$

The unit impulse response is shown in the figure 7. The **unit impulse response**, c(t) is an exponential decaying signal for positive values of 't' and it is zero for negative values of 't'.

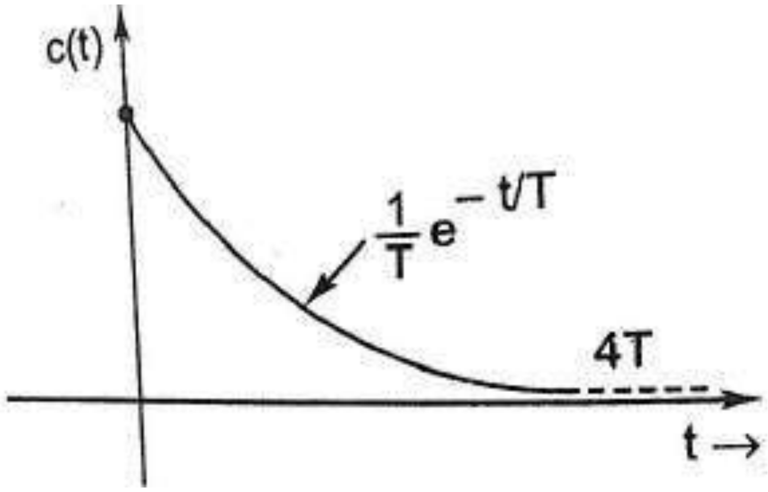


figure 7: Impulse Response of First Order System

<u>Step Response of First Order System</u>

For unit step signal R(s) = 1/s

Consider the equation (11), $C(s) = \frac{1}{1+Ts}R(s)$

$$\therefore \qquad C(s) = \frac{1}{1+Ts}\,\frac{1}{s} = \frac{1}{s(1+Ts)} \qquad (15)$$

Taking Partial fractions of Equ. (15)

$$C(s) = \frac{1}{s(1+Ts)} = \frac{A}{s} + \frac{B}{1+Ts} \qquad (16)$$

$$\Rightarrow \frac{1}{s(1+Ts)} = \frac{A(1+Ts)+Bs}{s(1+Ts)}$$

$$\Rightarrow 1 = A\,(sT+1) + Bs \qquad (17)$$

By solving Equ. (17), we get

$$A = 1 \; ; \; B = -T$$

Substitute, A = 1 and B = −T in Equ. (16), we get

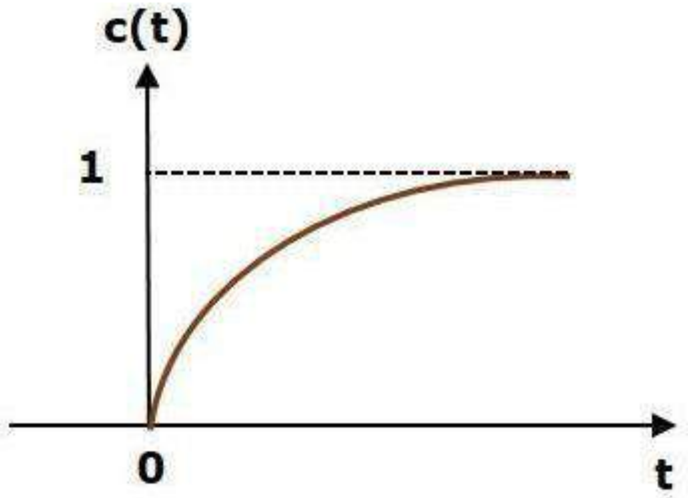
$$C(s) = \frac{1}{s} + \frac{(-T)}{1+Ts}$$

$$\Rightarrow C(s) = \frac{1}{s} - \frac{1}{\frac{1}{T}+s}$$

Apply inverse Laplace transform on both the sides.

$$c(t) = 1 - e^{-t/T} \quad (18)$$

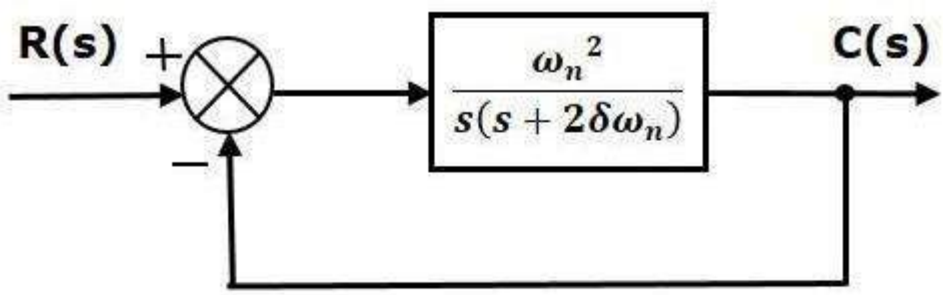
The following figure shows the unit step response.



The value of the **unit step response, c(t)** is zero at t = 0 and for all negative values of t. It is gradually increasing from zero value and finally reaches to one in steady state. So, the steady state value depends on the magnitude of the input.

Time Response of Second Order System

Consider the following block diagram of closed loop control system. Here, an open loop transfer function, $\omega_n^2 / s(s+2\zeta\omega_n)$ is connected with a unity negative feedback.



We know that the transfer function of the closed loop control system having unity negative feedback as

$$\frac{C(s)}{R(s)} = \frac{G(s)}{1+G(s)H(s)}$$

Substitute, G(s) = $\frac{\omega_n^2}{(s+2\zeta\omega_n)}$ in the above equation.

$$C(s)/R(s) = \frac{\frac{\omega_n^2}{(s+2\zeta\omega_n)}}{1+\frac{\omega_n^2}{(s+2\zeta\omega_n)}} = \frac{\omega_n^2}{s^2+2\zeta\omega_n s+\omega_n^2} \quad (19)$$

The power of 's' is two in the denominator term. Hence, the above transfer function is of the second order and the system is said to be the **second order system**.

The characteristic equation is -

$$s^2+2\zeta\omega_n s+\omega_n^2=0 \quad (20)$$

The roots of characteristic equation are -

$$S_1, S_2 = [-2\omega_n \zeta \pm \sqrt{(2\zeta\omega_n)^2 - 4\omega_n^2}\,]/2$$

- The two roots are imaginary when $\zeta = 0$.
- The two roots are real and equal when $\zeta = 1$.
- The two roots are real but not equal when $\zeta > 1$.
- The two roots are complex conjugate when $0 < \zeta < 1$.

We can write C(s) equation as,

$$C(s) = \frac{\omega_n^2}{s^2+2\zeta\omega_n s+\omega_n^2} R(s) \quad (21)$$

Step Response of Second Order System

Consider the unit step signal as an input to the second order system.

Laplace transform of the unit step signal is,

$$R(s)=1/s$$

We know the transfer function of the second order closed loop control system is,

$$\mathbf{C(s)/R(s)} = \frac{m_n^2}{s^2+2\zeta m_n s+m_n^2}$$

Case 1: $\zeta = 0$ (undamped system)

Substitute, δ=0 in the transfer function.

$$C(s)/R(s) = \frac{\omega_n^2}{s^2+\omega_n^2}$$

$$\Rightarrow C(s) = R(s)\,\frac{\omega_n^2}{s^2+\omega_n^2} \quad (22)$$

Substitute, R(s)=1/s in equation 22

$$C(s) = \frac{\omega_n^2}{(s^2+\omega_n^2)} \quad (23)$$

By using partial fraction the equation4 can be written as

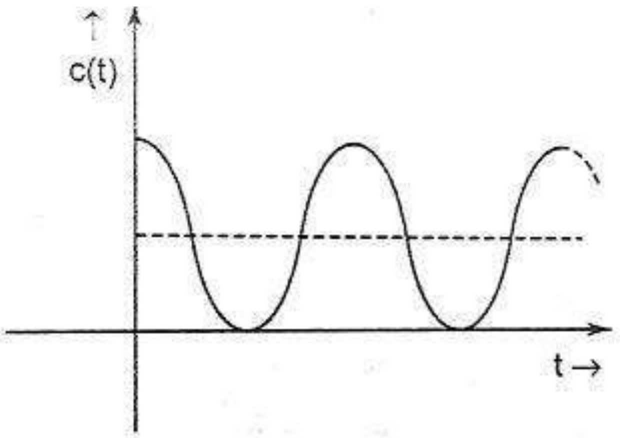
$$C(s) = \frac{A}{s} + \frac{Bs+C}{s^2+\omega_n^2} \quad (24)$$

After partial fraction A = 1; B = -1; C= 0 ∴ $C(s) = \frac{1}{s} - \frac{s}{s^2+\omega_n^2}$ (25)

Apply inverse Laplace transform on both the sides.

$$c(t) = 1-\cos(\omega_n t) \quad (26)$$

So, the unit step response of the second order system when $\zeta = 0$ will be a continuous time signal with constant amplitude and frequency. Since there is no damping with the time, this response does not die out with time. This response is known as undamped response as shown in the figure.



Undamped response of second order system with unit step input.

Case 2: $\zeta = 1$ (critically damped)

Substitute, $\zeta = 1$ in the transfer function.

$$C(s)/R(s) = \frac{\omega_n^2}{s^2+2\zeta\omega_n s+\omega_n^2} = \frac{\omega_n^2}{s^2+2\omega_n s+\omega_n^2}$$

$$\Rightarrow C(s) = R(s)\ \frac{\omega_n^2}{[s+\omega_n]^2} \quad (27)$$

Substitute, R(s)=1/s in equation 27

$$C(s) = \frac{\omega_n^2}{s[s+\omega_n]^2} \quad (28)$$

Do partial fractions of Equation 28

$$C(s)=\frac{\omega_n^2}{s[s+\omega_n]^2} = \frac{\boldsymbol{A}}{s} + \frac{\boldsymbol{B}}{s+\omega_n} + \frac{\boldsymbol{C}}{[s+\omega_n]^2} \qquad (29)$$

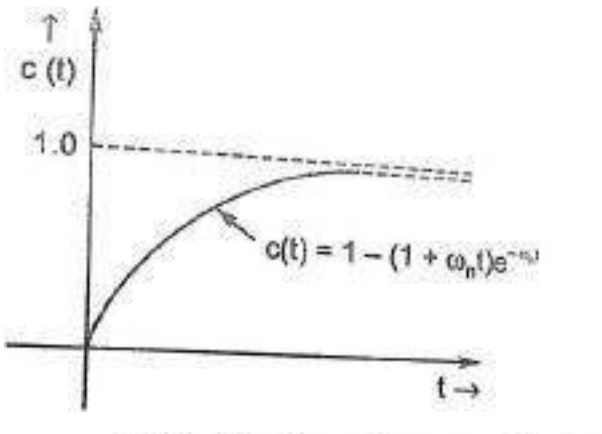
After simplifying, you will get the values of A, B and C as 1, −1 and and$-\omega_n$ respectively. Substitute these values in the above partial fraction expansion of C(s).

$$C(s)=\frac{\mathbf{1}}{s} + \frac{\mathbf{-1}}{s+\omega_n} + \frac{-\omega_n}{[s+\omega_n]^2} \qquad (30)$$

Apply inverse Laplace transform on both the sides of Eqs 30

$$c(t)=(1-e^{-\omega_n t} - \omega_n t e^{-\omega_n t}) = 1 - e^{-\omega_n t}(1 + \omega_n t) \qquad (31)$$

So, the unit step response of the second order system will try to reach the step input in steady state.



Critically damped system ($\xi = 1$).

Case 3: $0 < \zeta < 1$ (underdamped system)

From Equation (21) $C(s) = \frac{\omega_n^2}{s^2+2\zeta\omega_n s+\omega_n^2} R(s)$

Substitute, R(s)=1/s , Hence $C(s) = \frac{\omega_n^2}{(s^2+2\zeta\omega_n s+\omega_n^2)}$

Putting $s^2 + 2\zeta\omega_n s + \omega_n^2 = [s + \zeta\omega_n]^2 + \omega_n^2(1-\zeta^2)$, we get

$$C(s) = \frac{1}{s} \cdot \frac{\omega_n^2}{[s+\zeta\omega_n]^2 + \omega_n^2(1-\zeta^2)} \qquad (32)$$

Put $\omega_n^2(1-\zeta^2) = \omega_d^2$ and by doing partial fractions of Equation 32

$$C(s) = \frac{A}{s} + \frac{Bs+C}{[s+\zeta\omega_n]^2+\omega_d^2} \qquad (33)$$

After partial fractions we get

A= 1

B= -1

C = $-2\zeta\omega_n$

Putting the values of A, B,C in Equation 33, we have

$$C(s) = \frac{1}{s} - \frac{s+2\zeta\omega_n}{[s+\zeta\omega_n]^2 - \omega_d^2} = \frac{1}{s} - \frac{s+\zeta\omega_n}{[s+\zeta\omega_n]^2 + \omega_d^2} - \frac{\zeta\omega_n}{[s+\zeta\omega_n]^2+\omega_d^2}$$

$$= \frac{1}{s} - \frac{s+\zeta\omega_n}{[s+\zeta\omega_n]^2 + \omega_d^2} - \frac{\zeta}{\sqrt{1-\zeta^2}}\,\frac{\omega_n\sqrt{1-\zeta^2}}{[s+\zeta\omega_n]^2 + \omega_d^2}$$

$$= \frac{1}{s} - \frac{s+\zeta\omega_n}{[s+\zeta\omega_n]^2 + \omega_d^2} - \frac{\zeta}{\sqrt{1-\zeta^2}}\,\frac{\omega_d}{[s+\zeta\omega_n]^2 + \omega_d^2} \qquad (34)$$

Taking inverse laplace transform of Equation 34

$$C(t) = 1 - e^{-\zeta\omega_n t}\,\mathrm{Cos}\,(\omega_d \mathbf{t}) - \frac{\zeta}{\sqrt{1-\zeta^2}}\,e^{-\zeta\omega_n t}\,\mathrm{Sin}\,(\omega_d \mathbf{t})$$

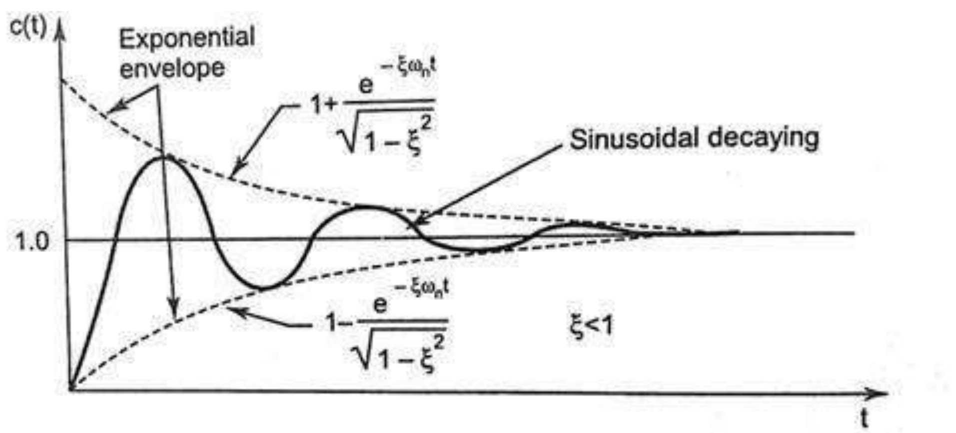
$$= 1 - \frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}}\left[\sqrt{1-\zeta^2}\,\mathrm{Cos}\,(\omega_d \mathbf{t}) - \zeta\,\mathrm{Sin}\,(\omega_d \mathbf{t})\right] \qquad (35)$$

Now putting $\zeta = \mathrm{Cos}\,\phi$; $\sqrt{1-\zeta^2} = \mathrm{Sin}\,\phi$ Since $\phi = \tan^{-1}\frac{\sqrt{1-\zeta^2}}{\zeta}$, hence equation 35 becomes

$$C(t) = 1 - \frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}}\left[\mathrm{Sin}\,\phi\;\mathrm{Cos}\,(\omega_d \mathbf{t}) - \mathrm{Cos}\,\phi\;\mathrm{Sin}\,(\omega_d \mathbf{t})\right]$$

$$= 1 - \frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}}\sin\,(\omega_d \mathbf{t} + \phi) \qquad (36)$$

Equation (36) represents the solution for **0** < ζ < **1** and it is represented in figure as given below.



Underdamped response of second order system to unit step input.

Case 4: **ζ** > 1

We can modify the denominator term of the second order transfer function as follows −

$$s^2 + 2\zeta\omega_n s + \omega_n^2 = [s + \zeta\omega_n]^2 - \omega_n^2(\zeta^2 - 1)$$

Hence from equation (32)

$$C(s) = \frac{1}{s}\cdot\frac{\omega_n^2}{[s+\zeta\omega_n]^2 - \omega_n^2(\zeta^2-1)} = \frac{1}{s}\cdot\frac{\omega_n^2}{(s+\zeta\omega_n+\omega_n\sqrt{\zeta^2-1})\,(s+\zeta\omega_n-\omega_n\sqrt{\zeta^2-1})} \qquad (37)$$

by doing partial fractions of Equation 37

$$C(s) = = \frac{A}{s} + \frac{B}{(s + \zeta\omega_n + \omega_n\sqrt{\zeta^2 - 1})} + \frac{C}{(s + \zeta\omega_n - \omega_n\sqrt{\zeta^2 - 1})} \tag{38}$$

After simplifying, you will get the values of

$$A = 1$$

$$B = \frac{1}{2}\frac{\zeta - \sqrt{\zeta^2 - 1)}}{\sqrt{\zeta^2 - 1)}}$$

$$C = -\frac{1}{2}\frac{\zeta + \sqrt{\zeta^2 - 1)}}{\sqrt{\zeta^2 - 1)}}$$

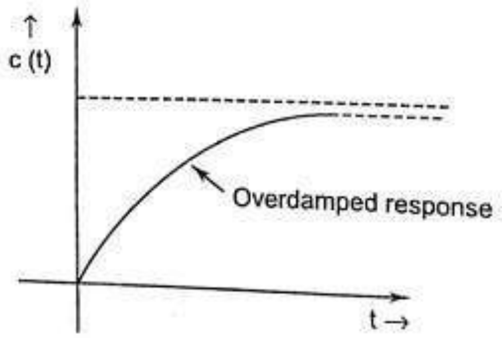
Substitute the value of A, B, C in equation (38)

$$C(s) = = \frac{1}{s} + \frac{1}{2}\frac{\zeta - \sqrt{\zeta^2 - 1)}}{\sqrt{\zeta^2 - 1)}}\frac{1}{(s + \zeta\omega_n + \omega_n\sqrt{\zeta^2 - 1})} - \frac{1}{2}\frac{\zeta + \sqrt{\zeta^2 - 1)}}{\sqrt{\zeta^2 - 1)}}\frac{1}{(s + \zeta\omega_n - \omega_n\sqrt{\zeta^2 - 1})} \tag{39}$$

Apply inverse Laplace transform of equation (29) we have

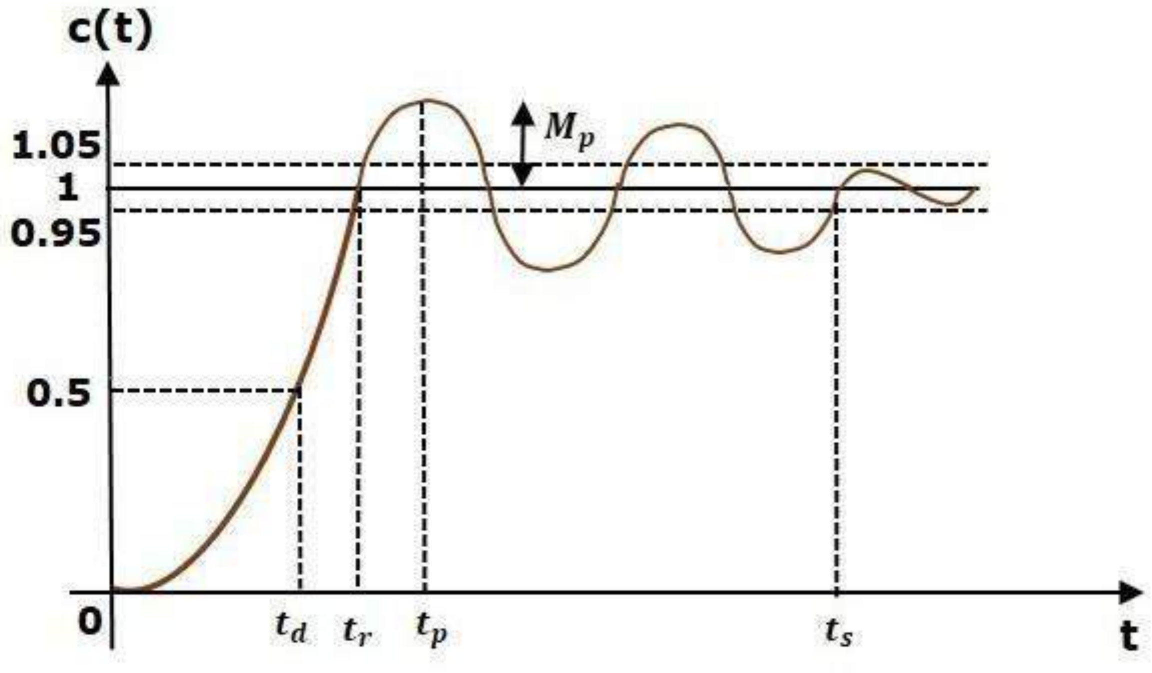
$$c(t) = 1 + \frac{\left(\xi - \sqrt{\xi^2 - 1}\right)}{2\sqrt{\xi^2 - 1}} e^{-\left(\xi\omega_n + \omega_n\sqrt{\xi^2 - 1}\right)t} - \frac{\xi + \sqrt{\xi^2 - 1}}{2\sqrt{\xi^2 - 1}} . e^{-\left(\xi\omega_n - \omega_n\sqrt{\xi^2 - 1}\right)t} \tag{40}$$

Equation (40) represents the solution for ζ > **1** and it is represented in figure as given below.



Time Domain Specifications

Let us discuss the time domain specifications of the second order system. The step response of the second order system for the under damped case is shown in the following figure.



All the time domain specifications are represented in this figure. The respse up to the settling time is known as transient response and the response after the settling time is known as steady state response.

1. Delay Time

It is the time required for the response to reach **50% of its final value** in first attempt. It is denoted by t_d.

Consider the step response of the second order system for $t \geq 0$, when 'ζ' lies between zero and one. From equation (36)

$$c(t) = 1 - \frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}} \sin(\omega_d \mathbf{t} + \phi)$$

The final value of the step response is one.

Therefore, at $t = t_d$, the value of the step response will be 0.5. Substitute these values in the above equation.

$$c(t_d) = 0.5 = 1 - \frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}} \sin(\omega_d t_d + \phi)$$

$$\Rightarrow \frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}} \sin(\omega_d t_d + \phi) = 0.5$$

By using linear approximation, you will get the **delay time t_d** as

$$t_d = (1+0.7\zeta)/\omega_n \qquad (40)$$

2. Rise Time

It is the time required for the response to rise from 10% to 90% of the final value for overdamped system and 0% to 100% of its final value for the **under-damped systems**. Rise time is denoted by $\mathbf{t_r}$.

At $t = t_r$, $c(t) = 1$

Hence from equation (36)

$$c(t) = 1 - \frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}}\sin(\omega_d \mathbf{t} + \phi)$$

$$c(t_r) = 1 = 1 - \frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}}\sin(\omega_d t + \phi)$$

$$\Rightarrow \frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}}\sin(\omega_d t_r + \phi) = 0$$

$$\Rightarrow \sin(\omega_d t_r + \phi) = 0$$

$$\Rightarrow \omega_d t_r + \phi = \pi$$

$$\Rightarrow t_r = (\pi - \phi)/\omega_d \tag{41}$$

From above equation, we can conclude that the rise time tr and the damped frequency ω_d are inversely proportional to each other.

3. Peak Time

It is the time required for the response to reach the **peak value** for the first time. It is denoted by t_p.

We know the step response of second order system for under-damped case is (from equation 36)

$$c(t) = 1 - \frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}}\sin(\omega_d \mathbf{t} + \phi)$$

At $t = t_p$, the first derivate of the response is zero. Hence

$$\frac{d(t_p)}{dt} = 0$$

$$\Rightarrow 0 - \frac{e^{-\zeta\omega_n t_p}(-\zeta\omega_n)in(\omega_d t_p + \phi)}{\sqrt{1-\zeta^2}} - \frac{e^{-\zeta\omega_n t_p}\omega_d\cos(\omega_d t_p + \phi)}{\sqrt{1-\zeta^2}} = 0$$

$$\Rightarrow \frac{e^{-\zeta\omega_n t}(-\zeta\omega_n)sin(\omega_d t_p + \phi)}{\sqrt{1-\zeta^2}} + \frac{e^{-\zeta\omega_n t_p}\omega_d\cos(\omega_d t_p + \phi)}{\sqrt{1-\zeta^2}} = 0$$

$$\Rightarrow \zeta\omega_n si(\omega_d t_p + \phi) = \omega_d\cos(\omega_d t_p + \phi)$$

$$\Rightarrow \tan(\omega_d t_p + \phi) = \omega_d / \zeta\omega_n$$

By putting $\omega_d = \omega_n\sqrt{1-\zeta^2}$ and $\phi = \tan^{-1}\frac{\sqrt{1-\zeta^2}}{\zeta}$

$$\tan(\omega_n\sqrt{1-\zeta^2}t_p + \phi) = \frac{\omega_n\sqrt{1-\zeta^2}}{\zeta\omega_n} = \tan\phi$$

or $\omega_n\sqrt{1-\zeta^2}t_p + \phi = n\pi + \phi$

for $n = 1$

$$t_p = \frac{\pi}{\omega_n\sqrt{1-\zeta^2}} \tag{42}$$

4. Peak Overshoot

Peak overshoot M_p is defined as the deviation of the response at peak time from the final value of response. It is also called the **maximum overshoot**.

Mathematically, we can write it as

$$M_p = c(t_p) - c(\infty)$$

Where,

$c(t_p)$ is the peak value of the response.

$c(\infty)$ is the final (steady state) value of the response.

From equation (36) $\quad c(t) = 1 - \frac{e^{-\zeta\omega_n t}}{\sqrt{1-\zeta^2}}\sin(\omega_d t + \phi)$

Put $t = t_p = \frac{\pi}{\omega_n\sqrt{1-\zeta^2}}$ and $\omega_d = \omega_n\sqrt{1-\zeta^2}$

∴ $$c(t_p) = 1 - \frac{e^{-\zeta\omega_n t_p}}{\sqrt{1-\zeta^2}}\sin(\omega_n\sqrt{1-\zeta^2}\,t_p + \phi)$$

$$= 1 - \frac{e^{-\zeta\omega_n\frac{\pi}{\omega_n\sqrt{1-\zeta^2}}}}{\sqrt{1-\zeta^2}}\sin\left(\omega_n\sqrt{1-\zeta^2}\,\frac{\pi}{\omega_n\sqrt{1-\zeta^2}} + \phi\right)$$

$$= 1 - \frac{e^{-\frac{\zeta\pi}{\sqrt{1-\zeta^2}}}}{\sqrt{1-\zeta^2}}\sin(\pi + \phi) = 1 + \frac{e^{-\frac{\zeta\pi}{\sqrt{1-\zeta^2}}}}{\sqrt{1-\zeta^2}}\sin\phi$$

$$= 1 + \frac{e^{-\frac{\zeta\pi}{\sqrt{1-\zeta^2}}}}{\sqrt{1-\zeta^2}}\sqrt{1-\zeta^2} \qquad \text{since } \sin\phi = \sqrt{1-\zeta^2}$$

∴ $$c(t_p) = 1 + e^{-\frac{\zeta\pi}{\sqrt{1-\zeta^2}}}$$

∴ $$\% M_P = \frac{c(t_p) - c(\infty)}{c(\infty)} \times 100 = \frac{1 + e^{-\frac{\zeta\pi}{\sqrt{1-\zeta^2}}} - 1}{1} \times 100$$

Or $$\% M_P = e^{-\frac{\zeta\pi}{\sqrt{1-\zeta^2}}} \times 100 \qquad (43)$$

5. Settling time

It is the time required for the response to reach the steady state and stay within the specified tolerance bands around the final value. In general, the tolerance bands are 2% and 5%. The settling time is denoted by t_s.

As seen from equation 26 the time constant of the exponential envelope is $T = 1/\zeta\omega_n$.

The settling time of the second order system for 2% tolerance band is appx. Four times the time constant T i.e.

$$t_s = 4/\zeta\omega_n = 4T$$

The settling time for 5% tolerance band is -

$$t_s = 3/\zeta\omega_n = 3T$$

Steady State Error

It indicate the error between the actual output and the desired output as t tends to infinity i.e.

$$e_{ss} = \lim_{t\to 0}[r(t) - c(t)] = \lim_{t\to 0} r(t) - \lim_{t\to 0} c(t)$$

$$= \lim_{t\to 0} 1 - \lim_{t\to 0}\left[1 - \frac{e^{-\zeta\omega nt}}{\sqrt{1-\zeta^2}}\sin(\omega_d \mathbf{t} + \phi)\right] = 1\text{-}1 = 0$$

Thus second order system has zero steady state error to unit step input.

Example

Let us now find the time domain specifications of a control system having the closed loop transfer function $4s^2 + 2s + 4$ when the unit step signal is applied as an input to this control system.

We know that the standard form of the transfer function of the second order closed loop control system as

$C(s)/R(s) = \omega_n^2 / s^2 + 2\zeta\omega_n s + \omega_n^2$

By equating these two transfer functions, we will get the un-damped natural frequency ω_n as 2 rad/sec and the damping ratio ζ as 0.5.

We know the formula for damped frequency ω_d as

$$\omega_d = \omega_n\sqrt{1-\zeta^2}$$

Substitute ω_n and δ values in the above formula.

$$\Rightarrow \omega_d = 2\sqrt{1-0.5^2}$$

$$\Rightarrow \omega_d = 1.732\text{rad/sec}$$

Substitute, δ value in following relation

$$\Phi = \cos^{-1}\zeta$$

$$\Rightarrow \phi = \cos^{-1}(0.5) = \pi/3\text{rad}$$

Substitute the above necessary values in the formula of each time domain specification and simplify in order to get the values of time domain specifications for given transfer function.

Steady State Error Analysis

The deviation of the output of control system from desired response during steady state is known as **steady state error**. It is represented as e_{ss}. We can find steady state error using the final value theorem as follows.

$$e_{ss} = \lim_{t\to\infty} e(t) = \lim_{s\to 0} sE(s)$$

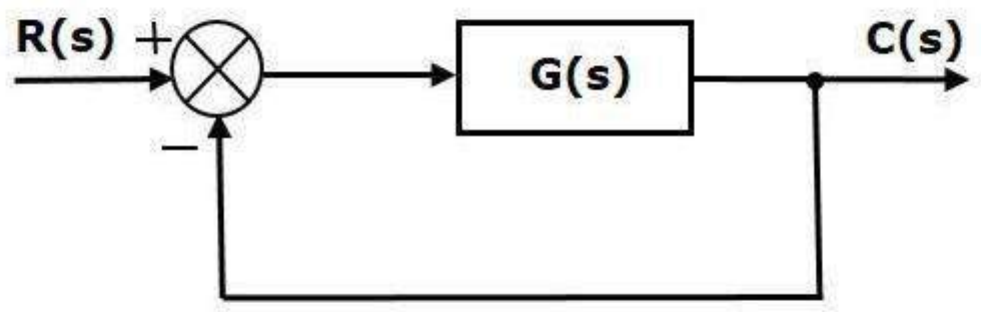
Where,

E(s) is the Laplace transform of the error signal, e(t)

Let us discuss how to find steady state errors for unity feedback and non-unity feedback control systems one by one.

Steady State Errors for Unity Feedback Systems

Consider the following block diagram of closed loop control system, which is having unity negative feedback.



Where,

- R(s) is the Laplace transform of the reference Input signal r(t)
- C(s) is the Laplace transform of the output signal c(t)

We know the transfer function of the unity negative feedback closed loop control system as

$$C(s)/R(s) = G(s)/1+G(s)$$

$$\Rightarrow C(s) = R(s)G(s)/1+G(s)$$

The output of the summing point is -

$$E(s) = R(s)-B(s) = R(s) - C(s)H(s) = R(s) - E(s)G(s)H(s)$$

$$\Rightarrow E(s)[1+ G(s)H(s)] = R(s)$$

$$\Rightarrow E(s)=R(s)/ [1+ G(s)H(s)]$$

For H(s) = 1 $\quad E(s)=R(s)/ [1+ G(s)]$

Substitute E(s) value in the steady state error formula

$$e_{ss} = \lim_{s \to 0} sR(s)/1 + G(s)$$

The following table shows the steady state errors and the error constants for standard input signals like unit step, unit ramp & unit parabolic signals.

Input signal	Steady state error e_{ss}	Error constant
unit step signal	1/1+kp	$K_p = \lim_{s \to 0} G(s)$
unit ramp signal	1/Kv	$K_v = \lim_{s \to 0} sG(s)$
unit parabolic signal	1/Ka	$K_a = \lim_{s \to 0} s2G(s)$

Where Kp, Kv and Ka are position error constant, velocity error constant and acceleration error constant respectively.

Note –

1. If any of the above input signals has the amplitude other than unity, then multiply corresponding steady state error with that amplitude.

2. We can't define the steady state error for the unit impulse signal because, it exists only at origin. So, we can't compare the impulse response with the unit impulse input as **t** denotes infinity.

Example

Let us find the steady state error for an input signal $r(t)=(5+2t+t^2/2)u(t)$ of unity negative feedback control system with $G(s)=5(s+4)/s^2(s+1)(s+20)$

The given input signal is a combination of three signals step, ramp and parabolic. The following table shows the error constants and steady state error values for these three signals.

Input signal	Error constant	Steady state error
$r_1(t)=5u(t)$	$K_p=\lim_{s \to 0} G(s)=\infty$	$e_{ss1}=5/1+kp = 0$
$r_2(t)=2tu(t)$	$K_v=\lim_{s \to 0} sG(s)=\infty$	$e_{ss2}=2/K_v=0$
$r_3(t)=t^2/2\, u(t)$	$K_a=\lim_{s \to 0} s^2 G(s)=1$	$e_{ss3}=1/ka=1$

We will get the overall steady state error, by adding the above three steady state errors.

$$e_{ss}= e_{ss1}+ e_{ss2}+ e_{ss3}$$

$$\Rightarrow e_{ss} = 0+0+1=1$$

Therefore, we got the steady state error e_{ss} as **1** for this example.

Steady state error for different types of input for Type-0, Type- 1and Type-2 systems.

Input	'Type-0' system			'Type-1' system		'Type-2' system	
	$\mathcal{L}$ (input)	Static error coefficient	steady state error e_{ss}	Static error coefficient	Steady state error e_{ss}	Static error coefficient	e_{ss}
Step input	$\frac{A}{s}$	$K_p = K$	$\frac{A}{1+K}$	$K_p = \infty$	0	$K_p = \infty$	0
Ramp input	$\frac{A}{s^2}$	$K_v = 0$	∞	$K_v = K$	$\frac{1}{K}$	$K_v = \infty$	0
Parabolic input	$\frac{A}{s^3}$	$K_a = 0$	∞	$K_a = 0$	∞	$K_v = K$	$\frac{1}{K}$

Effect of adding poles and zero to transfer function

Effect of addition of pole to transfer function:

1) As the pole moves towards the origin in s plane, the rise time increases and the maximum overshoot decreases, as far as the overshoot is concerned, adding a pole to the closed loop transfer function has just the opposite effect to that of adding a pole to forward path transfer function as discussed in the last article.

2) The addition of left half pole tends to slow down the system response.

3) The effect of addition of pole becomes more pronounced as pole location drifts away from imaginary axis.

4) Addition of right half pole will make overall system response to be an unstable one.

Effect of addition of zero to transfer function:

1) Makes the system overall response faster.

2) Rise time, peak time, decreases but overshoot increases.

3) Addition of right half zeros means system response slower and system exhibits inverse response. Such systems are said to be non-minimum phase systems.

COURSE CONTENTS

1. ANALYSIS OF STABILITY BY ROOT LOCUS TECHNIQUE.

3.1 Root locus concept.

3. 2 Construction of root loci.

3. 3 Rules for construction of the root locus.

3. 4 Effect of adding poles and zeros to G(s) and H(s).

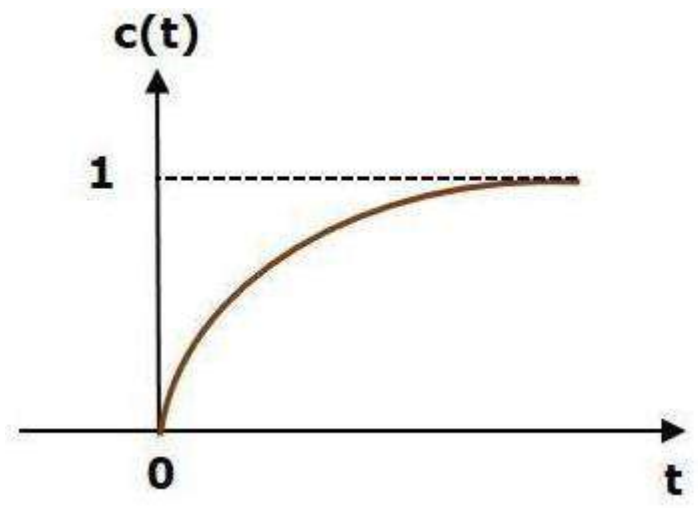
CHAPTER III

Analysis of stability by Root Locus Technique

What is Stability?

A system is said to be stable, if its output is under control. Otherwise, it is said to be unstable. A **stable system** produces a bounded output for a given bounded input.

The following figure shows the response of a stable system.



This is the response of first order control system for unit step input. This response has the values between 0 and 1. So, it is bounded output. We know that the unit step signal has the value of one for all positive values of **t** including zero. So, it is bounded input. Therefore, the first order control system is stable since both the input and the output are bounded.

Types of Systems based on Stability

We can classify the systems based on stability as follows.

- Absolutely stable system
- Conditionally stable system
- Marginally stable system

Absolutely Stable System

If the system is stable for all the range of system component values, then it is known as the **absolutely stable system**. The open loop control system is absolutely stable if all the poles of the open loop transfer function present in left half of **'s' plane**. Similarly, the closed loop control system is absolutely stable if all the poles of the closed loop transfer function present in the left half of the 's' plane.

Conditionally Stable System

If the system is stable for a certain range of system component values, then it is known as **conditionally stable system**.

Marginally Stable System

If the system is stable by producing an output signal with constant amplitude and constant frequency of oscillations for bounded input, then it is known as **marginally stable system**. The open loop control system is marginally stable if any two poles of the open loop transfer function is present on the imaginary axis. Similarly, the closed loop control system is marginally stable if any two poles of the closed loop transfer function is present on the imaginary axis.

Condition for stability

Let us consider a transfer function of a closed loop system:

$$\frac{(s)}{(s)} = \frac{a_0\, s^m + a_1\, s^{m-1} + a_2 s^{m-2} + \ldots + a_{m-1}\, s^1\, + a_m s^0}{a_0\, s^n + a_1\, s^{n-1} + a_2\, s^{n-2} + \ldots + a_{n-1}\, s^1\, + a_n\, s^0};$$

Here the characteristics Equation : $a_0\, s^n + a_1\, s^{n-1} + a_2\, s^{n-2} + \ldots + a_{n-1}\, s^1 + a_n\, s^0 = 0$

Necessary and sufficient conditions for stability:

1. All the coefficients of the ch. Equation should have same sign.

2. There should be no missing term.

Routh-Hurwitz Stability Criterion

This criterion is based on ordering the coefficients of the characteristics equation into an array called Routh's array. The Routh's array is formed as follows.

Follow this procedure for forming the Routh table.

- Fill the first two rows of the Routh array with the coefficients of the characteristic polynomial as mentioned in the table. Start with the coefficient of s_n and continue up to the coefficient of s_0.
- Fill the remaining rows of the Routh array with the elements as mentioned in the table . Continue this process till you get the first column element of **row** s_0 is a_n. Here, a_n is the coefficient of s_0 in the characteristic polynomial.

Note − If any row elements of the Routh table have some common factor, then you can divide the row elements with that factor for the simplification will be easy.

Consider the characteristic equation of the order 'n' is -

$a_0\, s^n + a_1\, s^{n-1} + a_2\, s^{n-2} + \ldots + a_{n-1}\, s^1 + a_n\, s^0 = 0$

s^n	a_0	a_2	a_4	a_6	…	…
s^{n-1}	a_1	a_3	a_5	a_7	…	…
s^{n-2}	b1=(a1a2−a3a0)/a1	b2=(a1a4−a5a0)/a1	b3=(a1a6−a7a0)/a1	…	…	…
s^{n-3}	c1=(b1a3−b2a1)/b1	c2=(b1a5−b3a1)/b1	⁝⁝			

⁝	⁝	⁝	⁝			
S^1	⁝	⁝				
S^0	an					

Sufficient Condition for Routh-Hurwitz Stability

The sufficient condition is that all the elements of the first column of the Routh array should have the same sign. This means that all the elements of the first column of the Routh array should be either positive or negative.

Example

Let us find the stability of the control system having characteristic equation,

$S^4+3s^3+3s^2+2s+1=0$

Step 1 − Verify the necessary condition for the Routh-Hurwitz stability.

All the coefficients of the characteristic polynomial, $S^4+3s^3+3s^2+2s+1$ are positive. So, the control system satisfies the necessary condition.

Step 2 − Form the Routh array for the given characteristic polynomial.

S^4	1	3	1
S^3	3	2	
S^2	$\frac{(3\times3)-(2\times1)}{3} = 7/3$	$\frac{(3\times1)-(0\times1)}{3} =1$	
S^1	$\frac{(7/3\times2)-(1\times3)}{7/3} =5/7$		
S^0	1		

Step 3 − Verify the sufficient condition for the Routh-Hurwitz stability.

All the elements of the first column of the Routh array are positive. There is no sign change in the first column of the Routh array. So, the control system is stable.

Special Cases of Routh Array

The two special cases are −

- The first element of any row of the Routh array is zero.
- All the elements of any row of the Routh array are zero.

Let us now discuss how to overcome the difficulty in these two cases, one by one.

First Element of any row of the Routh array is zero

If any row of the Routh array contains only the first element as zero and at least one of the remaining elements have non-zero value, then replace the first element with a small positive integer, ϵ. And then continue the process of completing the Routh table. Now, find the number of sign changes in the first column of the Routh table by substituting ϵ tends to zero.

Example

Let us find the stability of the control system having characteristic equation,

$S^4+2s^3+s^2+2s+1=0$

Step 1 − Verify the necessary condition for the Routh-Hurwitz stability.

All the coefficients of the characteristic polynomial, $S^4+2s^3+s^2+2s+1$are positive. So, the control system satisfied the necessary condition.

Step 2 − Form the Routh array for the given characteristic polynomial.

S^4	1	1	1
S^3	2	2	
S^2	0	1	
S^1			
S^0			

Special case (i) − Only the first element of row s^2 is zero. So, replace it by ϵ and continue the process of completing the Routh table

.

s4	1	1	1
s3	1	1	
s2	ϵ	1	
s1	[(ϵ×1)−(1×1])/ϵ = (ϵ−1)/ϵ		
s0	1		

Step 3 − Verify the sufficient condition for the Routh-Hurwitz stability.

As ϵ tends to zero, the Routh table becomes like this.

s4	1	1	1
s3	1	1	
s2	0	1	

s1	-∞		
s0	1		

There are two sign changes in the first column of Routh table. Hence, the control system is unstable.

All the Elements of any row of the Routh array are zero

In this case, follow these two steps −

- Write the auxilary equation, A(s) of the row, which is just above the row of zeros.
- Differentiate the auxiliary equation, A(s) with respect to s. Fill the row of zeros with these coefficients.

Example

Let us find the stability of the control system having characteristic equation,

$S^5+3s^4+s^3+3s^2+s+3=0$

Step 1 − Verify the necessary condition for the Routh-Hurwitz stability.

All the coefficients of the given characteristic polynomial are positive. So, the control system satisfied the necessary condition.

Step 2 − Form the Routh array for the given characteristic polynomial.

s5	1	1	1
s4	$\cancel{3}$ 1	$\cancel{3}$ 1	$\cancel{3}$ 1
s3	0	0	
s2			
s1			
s0			

The row s4 elements have the common factor of 3. So, all these elements are divided by 3.
Special case (ii) − All the elements of row s3 are zero. So, write the auxiliary equation, A(s) of the row s4.

$A(s)=s^4+s^2+1$

Differentiate the above equation with respect to s.

$\frac{dA(s)}{ds}=4s3+2s$

Place these coefficients in row s^3.

s5	1	1	1

s4	1	1	1
s3	4	2	
s2	0.5	1	
s1	−3		
s0	1		

Step 3 − Verify the sufficient condition for the Routh-Hurwitz stability.

There are two sign changes in the first column of Routh table. Hence, the control system is unstable.

In the Routh-Hurwitz stability criterion, we can know whether the closed loop poles are in on left half of the 's' plane or on the right half of the 's' plane or on an imaginary axis. So, we can't find the nature of the control system. To overcome this limitation, there is a technique known as the root locus. We will discuss this technique in the next two chapters.

Root Locus

In the root locus diagram, we can observe the path of the closed loop poles. Hence, we can identify the nature of the control system. In this technique, we will use an open loop transfer function to know the stability of the closed loop control system.

The Root locus is the locus of the roots of the characteristic equation by varying system gain K from zero to infinity.

Angle Condition and Magnitude Condition

The points on the root locus branches satisfy the angle condition. So, the angle condition is used to know whether the point exist on root locus branch or not. We can find the value of K for the points on the root locus branches by using magnitude condition. So, we can use the magnitude condition for the points, and this satisfies the angle condition.

Characteristic equation of closed loop control system is

$1+G(s)H(s)=0$

$\Rightarrow G(s)H(s) = -1 + j0$

The **phase angle** of G(s)H(s) is
$\angle G(s)H(s)=\tan^{-1} 0/(-1)=(2n+1)\pi$

The **angle condition** is the point at which the angle of the open loop transfer function is an odd multiple of 180^0.

Magnitude of G(s)H(s) is -
$|G(s)H(s)| =\sqrt{(-1)^2 + 0^2} =1$

The magnitude condition is that the point (which satisfied the angle condition) at which the magnitude of the open loop transfer function is one.

Rules for Construction of Root Locus

Follow these rules for constructing a root locus.

Rule 1 − Locate the open loop poles and zeros in the ‘s’ plane.

Rule 2 − Find the number of root locus branches.

We know that the root locus branches start at the open loop poles and end either at open loop zeros or at ∞. So, the number of root locus branches **N** is equal to the number of finite open loop poles **P** or the number of finite open loop zeros **Z**, whichever is greater.

Mathematically, we can write the number of root locus branches **N** as

$N=P$ if $P \geq Z$
$N=Z$ if $P<Z$

Rule 3 − Identify and draw the **real axis root locus branches**.

A point or segment on the real axis lies on the root locus if the sum of open loop poles and open loop zeros to the right of this point or segment is odd.

Rule 4 − Find the centroid and the angle of asymptotes.

Asymptotes give the direction of these root locus branches.
Number of Asymptotes= P -Z
The intersection point of asymptotes on the real axis is known as **centroid**.

We can calculate the **centroid** $\boldsymbol{\sigma_A}$ by using this formula,

$$\boldsymbol{\sigma_A} = \frac{\sum \text{Real part of finite open loop poles} - \sum \text{Real part of finite open loop zeros}}{\text{P} - \text{Z}}$$

The formula for the angle of **asymptotes** is

$$\Phi_A = \frac{(2q+1)180}{P-Z}$$

Where,

$q=0,1,2,...,(P-Z-1)$

Rule 5 − Find the intersection points of root locus branches with an imaginary axis.

We can calculate the point at which the root locus branch intersects the imaginary axis and the value of **K** at that point by using the Routh array method

Rule 6 − Find Break-away and Break-in points.

- If there exists a real axis root locus branch between two open loop poles, then there will be a **break-away point** in between these two open loop poles.
- If there exists a real axis root locus branch between two open loop zeros, then there will be a **break-in point** in between these two open loop zeros.

Note − Break-away and break-in points exist only on the real axis root locus branches.

Follow these steps to find break-away and break-in points.

- Write K in terms of s from the characteristic equation $1+G(s)H(s)=0$

- Differentiate K with respect to s and make it equal to zero. Substitute these values of s in the above equation.
- The values of s for which the K value is positive are the **break points**.

Rule 7 − Find the angle of departure and the angle of arrival.

The Angle of departure and the angle of arrival can be calculated at complex conjugate open loop poles and complex conjugate open loop zeros respectively.

The formula for the **angle of departure** ϕ_d is

Φ_d = 180 –sum of the angles of vectors drawn to this pole to other poles + sum of the angles of vectors drawn to this pole to zeros

The formula for the **angle of arrival** φ_a is

Φ_a = 180 –sum of the angles of vectors drawn to this zero to other zeros + sum of the angles of vectors drawn to this zero to poles

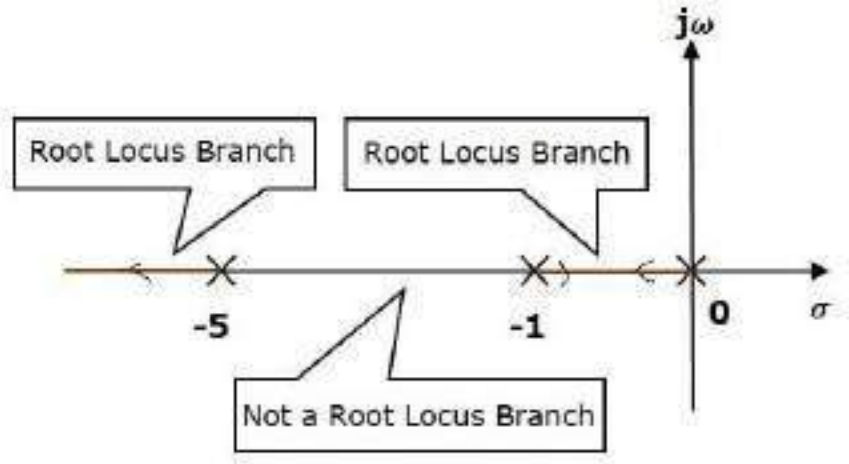
Example

Let us now draw the root locus of the control system having open loop transfer function

$G(s)H(s)= \frac{k}{s(s+1)(s+5)}$

Step 1 − The given open loop transfer function has three poles at s=0, s=−1 and s=−5. It doesn't have any zero. Therefore, the number of root locus branches is equal to the number of poles of the open loop transfer function.

N=P=3



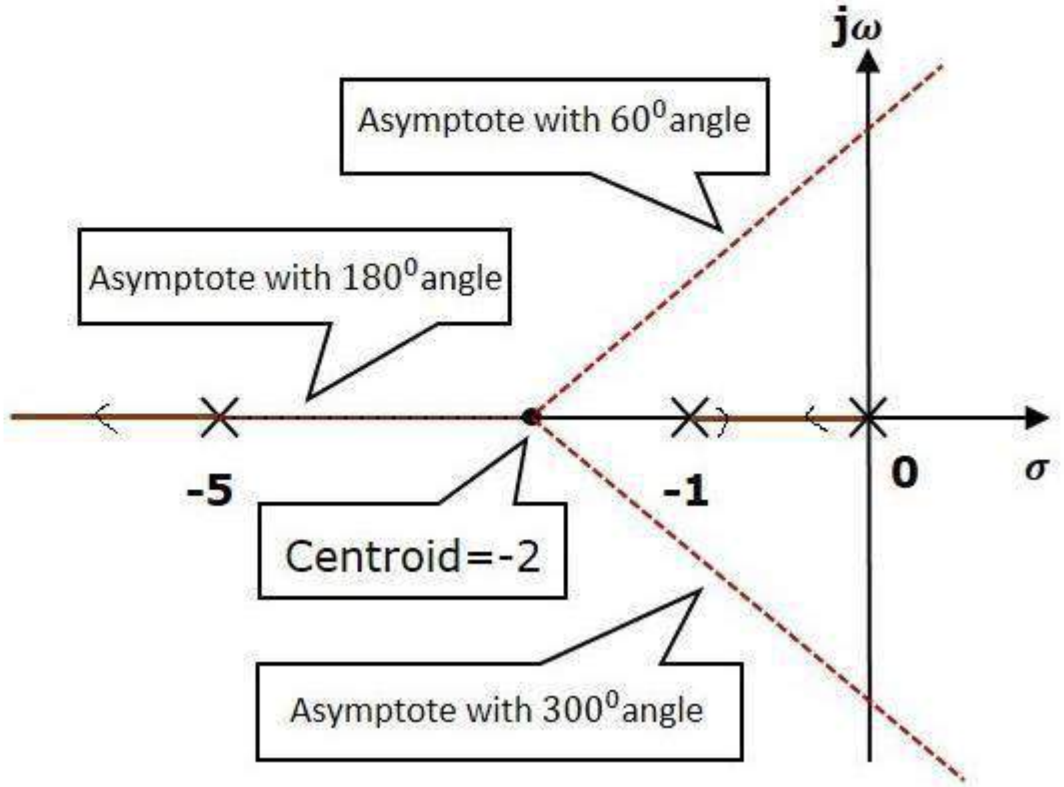
The three poles are located are shown in the above figure. The line segment between s=−1 and s=0 is one branch of root locus on real axis. And the other branch of the root locus on the real axis is the line segment to the left of s=−5 i.e in between -5 and ∞.

Step 2 − We will get the values of the centroid and the angle of asymptotes by using the given formulae.

Centroid $\boldsymbol{\sigma_A} = \frac{0-1-5}{3-0} = -2$

The angle of asymptotes $\Phi_A = \frac{(2q+1)180}{P-Z} = \frac{(2q+1)180}{3-0}$ for q= 0, 1, 2 angle of asymptotes are θ=60°,180° and 300°

The centroid and three asymptotes are shown in the following figure.



Step 3 − Since two asymptotes have the angles of 60° and 300°, two root locus branches intersect the imaginary axis. By using the Routh array method and special case(ii), the intersects of root locus branches to the imaginary axis can be found out as below

The characteristics equation of the given TF is 1+G(s)H(s)= 0

$$\text{Or } 1+\frac{k}{s(s+1)(s+5)} = 0$$

$$\text{Or } s^3+6s^2+5s+K = 0$$

Routh array

s3	1	5
s2	6	k
s1	$\frac{30-k}{6}$	0
s0	k	

For system stability the coefficient of Routh's array having positive and non zero value hence:

K > 0

$\frac{30-k}{6} > 0$ or k< 30

The range of K for which the system became stable is 0< k <30

At k = 30, the system auxiliary equation is

$6s^2 + 30= 0$

Or $s = \pm j\sqrt{5}$

Hence the root locus intersect the imaginary axis at $\pm j\sqrt{5}$

Step 4 − There will be one break-away point on the real axis root locus branch between the poles s=−1 and s=0. By following the procedure given for the calculation of break-away point,

The characteristics equation $s^3+6s^2+5s+K = 0$

Or $K= -(s^3+6s^2+5s)$

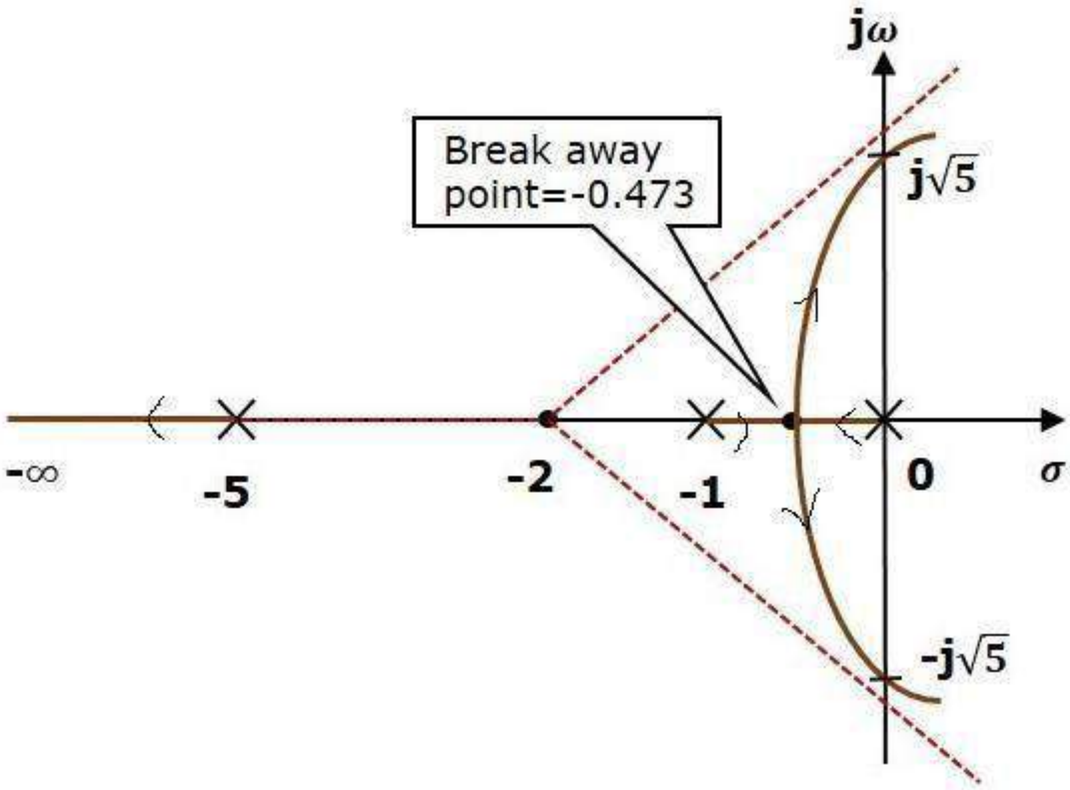
$\frac{dk}{ds} = 0$

Or $3s^2+12s+5=0$

The roots of s= -0.473, -3.52

Since breakaway point must lie between 0 and -1, it is clear that s=−0.473is actual breakaway point.

The root locus diagram for the given control system is shown in the following figure.



<u>Example</u> :- A feedback control system has open-loop transfer function

$$G(s)H(s) = \frac{K_a}{s(s+3)(S^2+2s+2)}$$

Draw root locus as K is varied from 0 to ∞ .

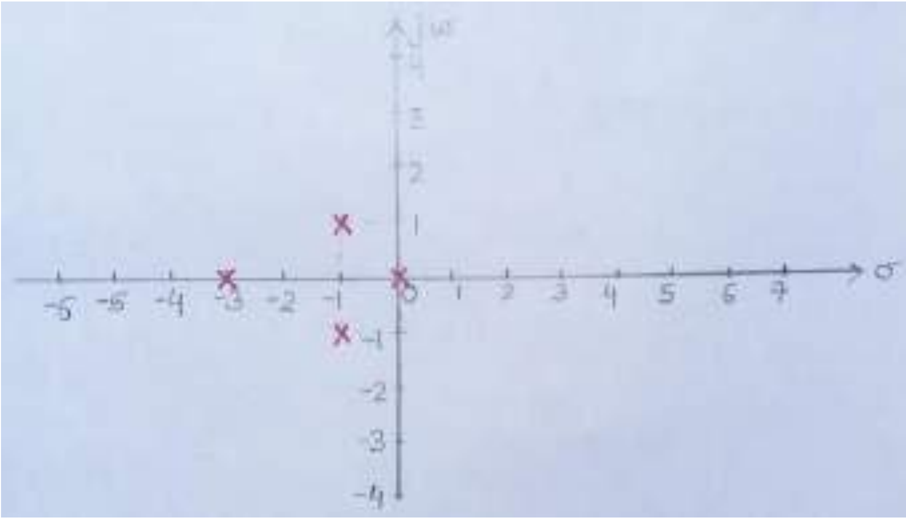
Solution:

<u>Step-1</u> :- Find OL poles and OL zeros from the OLTF

OL poles are S=0,-3, (-1+j1) and (-1-j1)

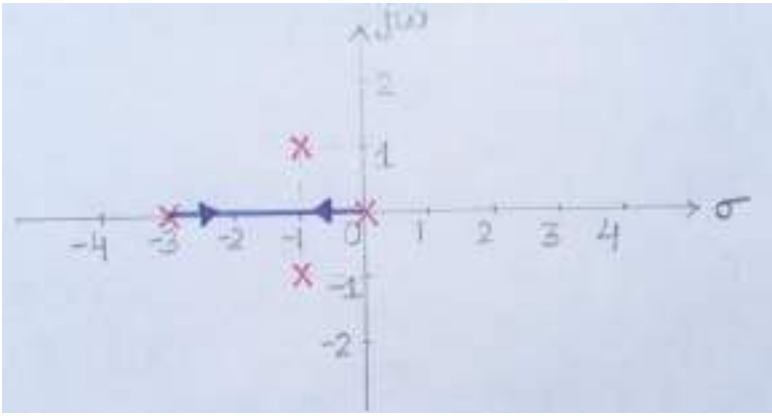
There are no finite OL zeros.

Mark OL pole with cross and OL zero with circle in S-plane as shown.



<u>Step-2</u> Find the parts of the real axis at which root locus lies.

A point on real axis lies on root locus if the number of OL poles+OL zeros on the real axis to the right of the point is odd.Hence the Root locus exist between s=0 and s= -3 in the real axis.



<u>Step-3</u> Number of root locus branches N = P= 4

<u>Step-4</u> Find number of asymptotes:

Number of asymptotes = P - Z = 4 (where P,Z = nos of open loop pole and zero)

<u>Step-</u>5Calculation for cetroid

$$\boldsymbol{\sigma_A} = \frac{\sum \text{Real part of finite open loop poles} - \sum \text{Real part of finite open loop zeros}}{\text{P} - \text{Z}}$$

$$= \frac{(0-3-1-1)-(0)}{4} = -1.25$$

<u>Step-</u>6 Calculation for asymptotic angle:

$\Phi_A = \frac{(2q+1)180}{P-Z}$ For q=0; $\Phi_A = \frac{180\,(0+1)}{4} = 45^0$

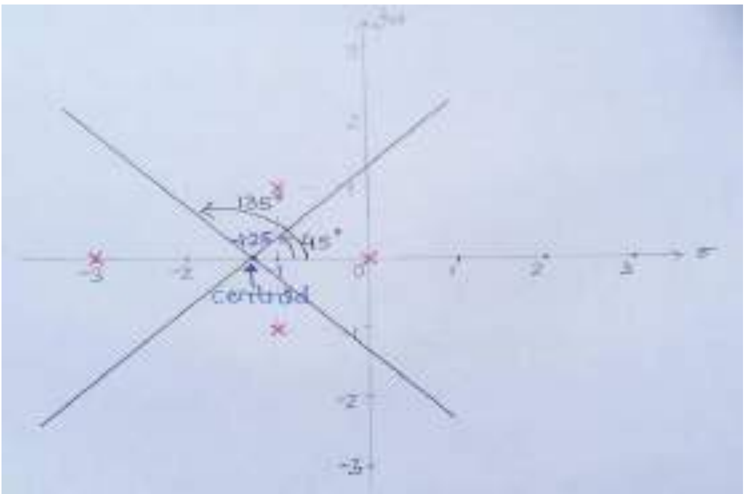
For q=1; $\Phi_A = \frac{180\,(2+1)}{4} = 135^0$

For q=2; $\Phi_A = 225^0$

$\Phi_A = 315^0$

For q=3;

So, from steps 2,3 and 4 , four asymptotes cut the real axis at -1.25 and make angles 45^0, 135^0, 225^0 and 315^0 , as shown below.



Step-7: Find the *breakaway points* (points at which two or more root locus branches meet)

Breakaway points are the solutions of $(dK_a/ds)=0$

The characteristic equation will be $S(S+3)(S^2+2S+2)+K_a=0$

From the characteristic equation,

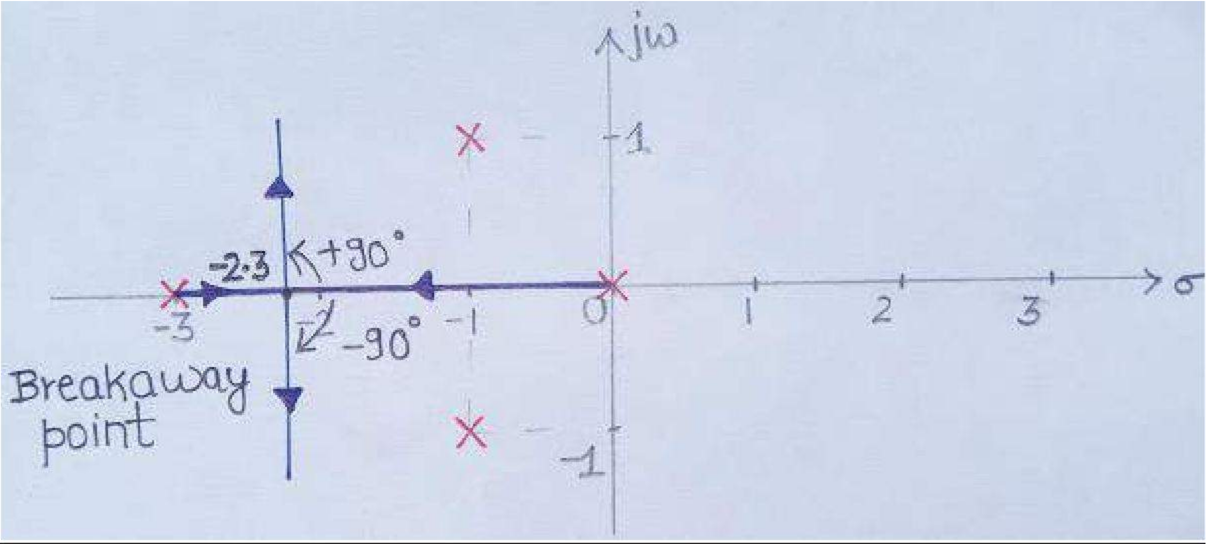
$K_a = -S(S+3)(S^2+2S+2) = -(S^4+5S^3+8S^2+6S)$

$$\therefore \frac{dK_a}{ds} = -4(S^3+3.75S^2+4S+1.5) = 0$$

We get, S = -2.3 , (-0.725±j0.365)

Not all values obtained as solutions of $(dK_a/ds)=0$ need to be necessarily the breakaway points. Out of the obtained s values only those values of S which satisfy angle condition are the actual breakaway points.

On checking angle condition we find that (-0.725±j0.365) do not satisfy it. Therefore, only S= -2.3 is the only breakaway point. So, the real axis from 0 to -3 contains root locus which breakdown at -2.3 as shown.



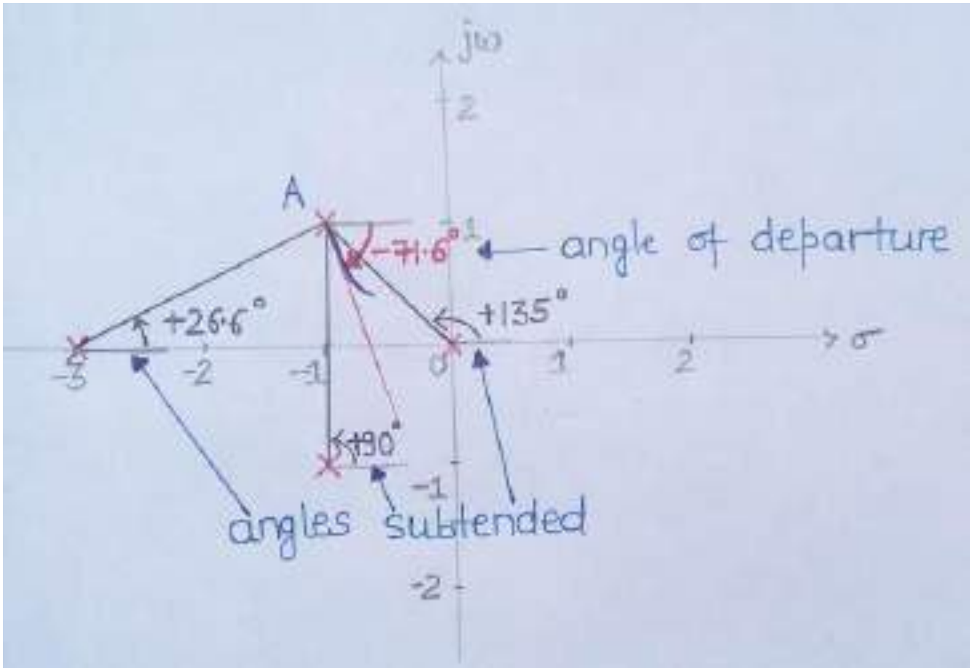
Step-8 :- Find *angles of departure* as there is a presence of pole in complex plane (angle which a root locus branch starting from an open loop pole, makes with a line parallel to the asymptotic line.

The formula for the **angle of departure** ϕ_d is

Φ_d = 180 –sum of the angles of vectors drawn to this pole to other poles + sum of the angles of vectors drawn to this pole to zeros

$$\text{Or } \Phi_d = 180 - (90^0 + 135^0 + 26.6^0) = -71.6^0$$

So, root locus branch starts from (-1+j1) at an angle -71.6^0 . Since root locus is always mirror image about real axis , therefore, root locus starts from (-1-j1) at $+71.6^0$.



Step-9 :- Find the points at which root locus branches intersect jw axis.

The characteristic equation will be $S(S+3)(S^2+2S+2)+K_a = 0$

Or $S^4+5S^3+8S^2+6S+K_a=0$, Make rouths array;

S^4	1	8	K_a
S^3	5	6	
S^2	$\frac{(5\times8)-(6\times1)}{5} = 6.8$	K_a	
S^1	$\frac{(6.8\times6)-(Ka\times5)}{6.8}$		
S^0	K_a		

For the system to be stable all the coefficient of the first column of the Routh's array having positive and non zero value. Hence for system stability

$$K_a > 0$$

$$\frac{(6.8\times6)-(Ka\times5)}{6.8} > 0$$

$$\text{Or } 0 < K_a < 8.16$$

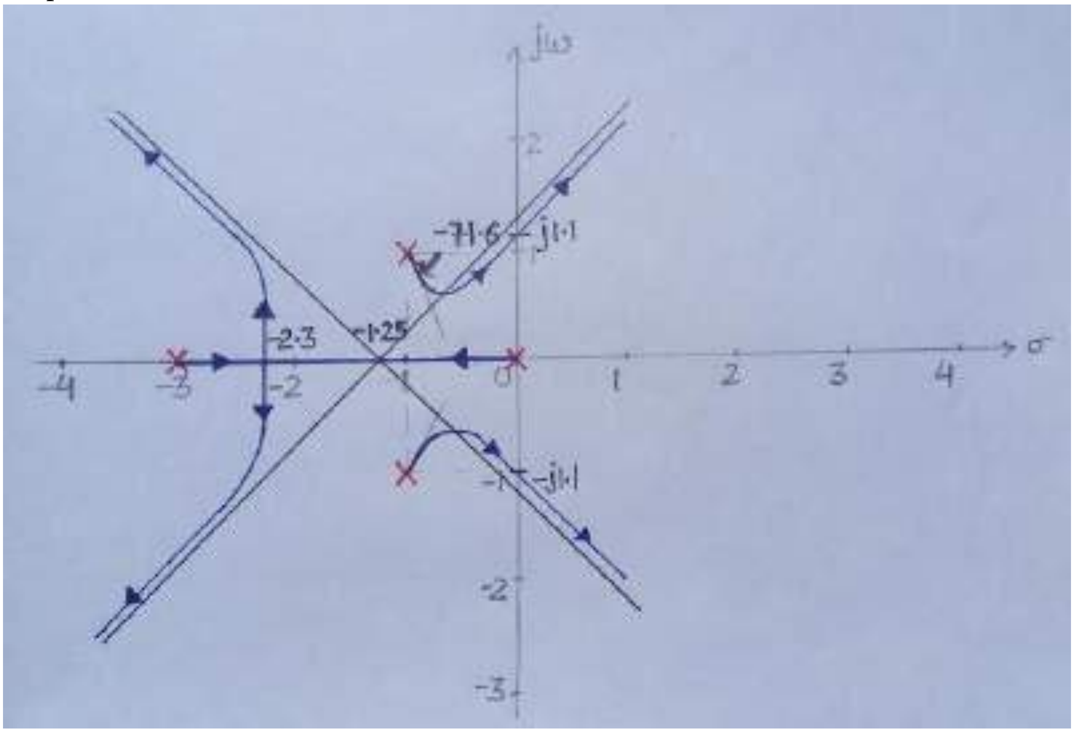
For $K_a = 8.16$ the Auxiliary equation is $6.8s^2 + 8.16 = 0$

$$\text{Or } s^2 = -1.2$$

$$\text{Or } s = \pm j1.1$$

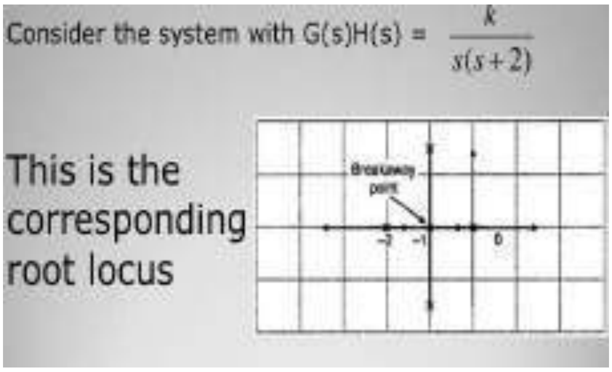
The points of intersection comes out to be +j1.1 and –j1.1

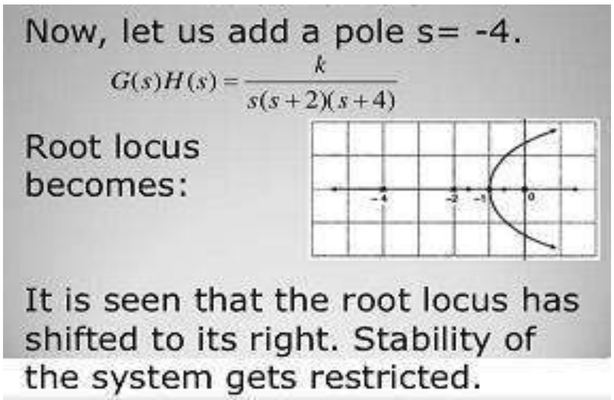
The complete root locus is shown below.

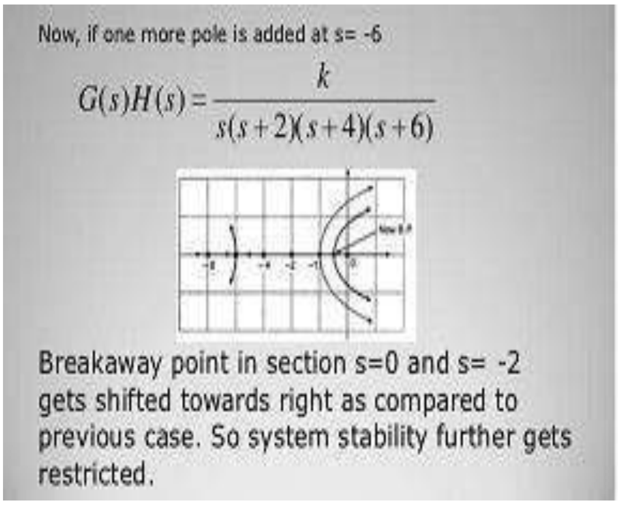


Effects of Adding Open Loop Poles and Zeros on Root Locus

<table>
<tr><th>Effect of addition of open loop pole</th><th>Effect of addition of open loop zero</th></tr>
<tr><td>







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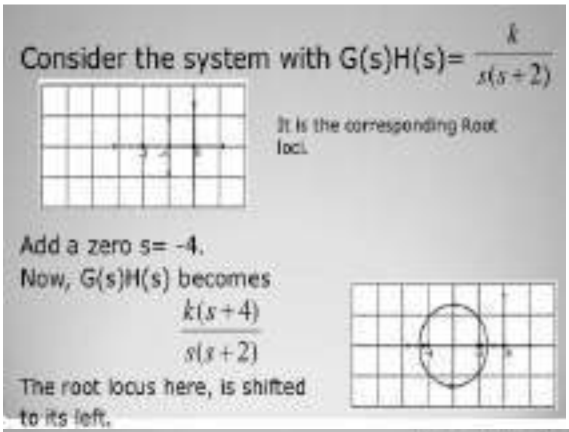

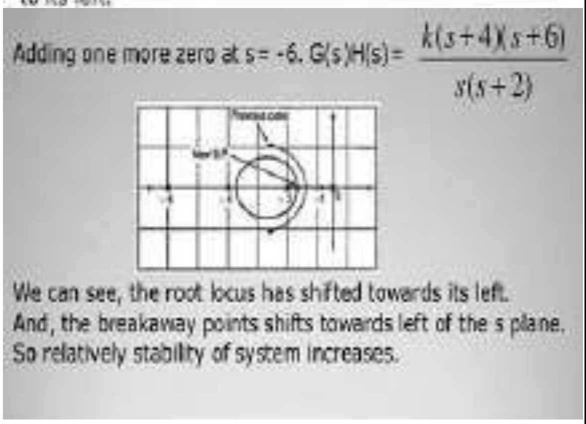

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</table>

➢The root locus changes its nature and shifts towards imaginary axis. ➢The system becomes oscillatory. ➢Gain Margin enhances relatively, thus stability decreases. ➢Range of k reduces. ➢Settling time increases.	➢The root locus changes its nature and shifts to left away from imaginary axis. ➢Relative stability of system increases. ➢System becomes less oscillatory. ➢Gain margin increases and so does the range of K. ➢Settling time decreases.

MODULE – IV
Frequency Response Analysis

Introduction:

The response of a system can be partitioned into both the transient response and the steady state response. We can find the transient response by using Fourier integrals. The steady state response of a system for an input sinusoidal signal is known as the **frequency response**. In this chapter, we will focus only on the steady state response.

If a sinusoidal signal is applied as an input to a Linear Time-Invariant (LTI) system, then it produces the steady state output, which is also a sinusoidal signal. The input and output sinusoidal signals have the same frequency, but different amplitudes and phase angles.

Let the input and output signal be −

$$r(t)=A\sin(\omega t) \tag{1}$$

$$c(t)=B\sin(\omega t+\phi) \tag{2}$$

The open loop transfer function will be −

$$G(s)=G(j\omega)$$

We can represent $G(j\omega)$ in terms of magnitude and phase as shown below.

$$G(j\omega)=|G(j\omega)|\angle G(j\omega)$$

The output signal is

$$c(t)=A|G(j\omega)|\sin(\omega t+\angle G(j\omega)) \tag{3}$$

- The **amplitude** of the output sinusoidal signal is obtained by multiplying the amplitude of the input sinusoidal signal and the magnitude of $G(j\omega)$ at ω.
- The **phase** of the output sinusoidal signal is obtained by adding the phase of the input sinusoidal signal and the phase of $G(j\omega)$ at ω

Where,

- **A** is the amplitude of the input sinusoidal signal.
- **Ω** is angular frequency of the input sinusoidal signal.

We can write, angular frequency ω as shown below.

$$\omega=2\pi f$$

Here, f is the frequency of the input sinusoidal signal. Similarly, you can follow the same procedure for closed loop control system.

Correlation between time and frequency response:

The frequency domain specifications are resonant peak, resonant frequency and bandwidth.

Consider the transfer function of the second order closed loop control system as,

$$T(s) = C(s)/R(s) = \omega_n^2/(s^2+2\zeta\omega_n s+\omega_n^2)$$

Substitute, s= jω in the above equation.

$$T(j\omega) = \omega_n^2 / (j\omega)^2+2\zeta\omega_n(j\omega)+\omega_n^2$$

$$\Rightarrow T(j\omega) = \frac{\omega_n^2}{-\omega^2+2j\zeta\omega\omega_n+\omega_n^2} = \frac{\omega_n^2}{\omega_n^2}\,\frac{1}{\left(-\frac{\omega^2}{\omega_n^2}\right)+2j\zeta\left(\frac{\omega}{\omega_n}\right)+1}$$

$$\Rightarrow T(j\omega) = \frac{1}{\left(1-\frac{\omega^2}{\omega_n^2}\right)+j2\left(\frac{\omega}{\omega_n}\right)} \qquad (4)$$

Let, $\frac{\omega}{\omega_n}$ **=u** Substitute this value in the above equation.

$$T(j\omega) = \frac{1}{(1-u^2)+j2\zeta u} \qquad (5)$$

Magnitude of T(jω) is -

$$M = |T(j\omega)| = \frac{1}{\sqrt{(1-u^2)^2+(2\zeta u)^2}} \qquad (6)$$

Phase of T(jω) is -

$$\angle T(j\omega) = -\tan^{-1}\frac{2\zeta u}{(1-u^2)} \qquad (7)$$

The steady-state output of the system for a sinusoidal input of unit magnitude and variable frequency ω is given by

$$C(t) = \frac{1}{\sqrt{(1-u^2)^2+(2\zeta u)^2}}\,\mathrm{Sin}\left(\omega t - \tan^{-1}\frac{2\zeta u}{(1-u^2)}\right) \qquad (8)$$

Resonant Frequency:

It is the frequency at which the magnitude of the frequency response has peak value for the first time. It is denoted by ω_r. At $\omega=\omega_r$ the first derivate of the magnitude of T(jω) is zero.

Differentiate M with respect to u.

$$\left.\frac{dM}{du}\right|_{u=u_r} = -\frac{1}{2}\,\frac{-4(1-u_r^2)u_r+8\zeta^2 u_r}{\left[(1-u_r^2)^2+(2\zeta u_r)^2\right]^{3/2}} = 0$$

$$\Rightarrow 4u_r^3 - 4u_r + 8\zeta^2 u_r = 0$$

$$\Rightarrow u_r = \sqrt{1-2\zeta^2} \qquad (9)$$

$$\text{i.e,}\quad \omega_r = \omega_n\sqrt{1-2\zeta^2} \qquad (10)$$

Resonant Peak:

It is the peak (maximum) value of the magnitude of T(jω). It is denoted by Mr.

At $u=u_r$, the Magnitude of T(jω) is -

$$M = |T(j\omega)| = \frac{1}{\sqrt{(1-u^2)^2 + (2\zeta u)^2}}$$

Substitute, $u_r = \sqrt{1 - 2\zeta^2}$ in the above equation, we get

$$M_r = \frac{1}{2\zeta\sqrt{1-\zeta^2}} \quad (11)$$

The phase angle of T(jω) at resonant frequency u_r obtained from equation 7 is

$$\Phi_r = -\tan^{-1} [\sqrt{1 - 2\zeta^2} / \zeta] \quad (12)$$

Resonant peak in frequency response corresponds to the peak overshoot in the time domain transient response for certain values of damping ratio . So, the resonant peak and peak overshoot are correlated to each other.

Bandwidth:

It is the range of frequencies over which, the magnitude of T(jω) drops to 70.7% (0.707) from its zero frequency value.

At ω=0, the value of u will be zero.

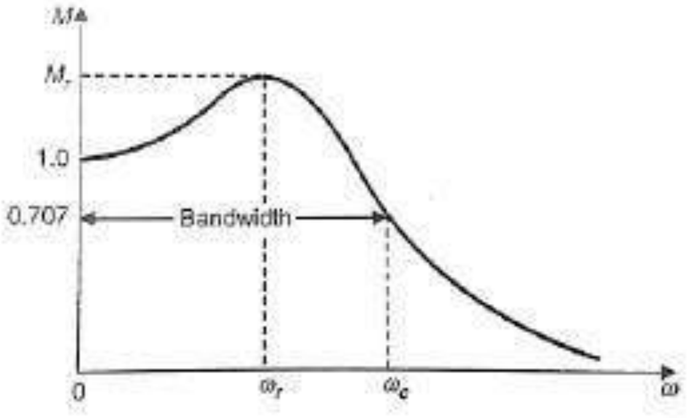
Substitute, u=0 in M, frm equation 6

$$M = \frac{1}{\sqrt{(1-0^2)^2 + (2\zeta 0)^2}} = 1$$

Therefore, the magnitude of T(jω) is one at ω=0.

At 3-dB frequency, the magnitude of T(jω) will be 70.7% of magnitude of T(jω) at ω=0.

i.e., at $\omega=\omega_b$, $M=0.707(1) = 1/\sqrt{2}$



Typical magnification curve of a feedback control system.

From Equation 6: $$M = \frac{1}{\sqrt{2}} = \frac{1}{\sqrt{(1-u_b^2)^2 + (2\zeta u_b)^2}}$$

$$\Rightarrow 2 = (1-u_b^2)^2 + (2\zeta)^2 u_b^2$$

Let, $u_b^2 = x$

$$\Rightarrow 2 = (1-x)^2 + (2\zeta)^2 x \Rightarrow x^2 + (4\zeta^2 - 2)x - 1 = 0$$

$$\Rightarrow x = \frac{-(4\zeta^2-2)\pm\sqrt{(4\zeta^2-2)^2-4}}{2}$$

Consider only the positive value of x.

$$x = \frac{-(4\zeta^2-2)+\sqrt{(4\zeta^2-2)^2-4}}{2}$$

or $$X = 1 - 2\zeta^2\sqrt{2 - 4\zeta^2 + 4\zeta^4}$$

Substitute, $x = u_b^2 = \omega_b^2/\omega_n^2$

$$\omega_b^2/\omega_n^2 = 1 - 2\zeta^2\sqrt{2 - 4\zeta^2 + 4\zeta^4}$$

$$\Rightarrow \omega_b = \omega_n\sqrt{1 - 2\zeta^2\sqrt{2 - 4\zeta^2 + 4\zeta^4}} \qquad (13)$$

Bandwidth ωb in the frequency response is inversely proportional to the rise time tr in the time domain transient response.

Bode Plots

Sinusoidal transfer function is graphically represented by Bode plot for determining the stability of the control system. Bode plot is a logarithmic plot and consists of two plots.

- A plot of the logarithmic (base 10) of magnitude (in decibel) Vs frequency in logarithmic scale i.e. log ω.
- A plot of Phase plot (ϕ) Vs frequency in logarithmic scale i.e. log ω.

In both the plots, x-axis represents angular frequency (logarithmic scale). Whereas, y-axis represents the magnitude (linear scale) of open loop transfer function in the magnitude plot and the phase angle (linear scale) of the open loop transfer function in the phase plot.

The **magnitude** of the open loop transfer function in dB (decibel) is -

$$M=20\log|G(j\omega)H(j\omega)| \qquad (1)$$

The **phase angle** of the open loop transfer function in degrees is -

$$\phi=\angle G(j\omega)H(j\omega) \qquad (2)$$

Note − the base of logarithm is 10.

Basic of Bode Plots

Let the generalised expression for open-loop transfer function of a system be given by:

$$G(s)H(s) = \frac{K\left[(1+sT_1)(1+sT_2)\ldots\ldots\right]\omega_n^2}{s^n\left[(1+sT_a)(1+sT_b)\ldots\ldots\right]\left[s^2+2\xi\omega_n s+\omega_n^2\right]} \qquad (3)$$

where $K, T_1, T_2, \ldots, T_a, T_b, \ldots, \xi, \omega_n$ are all real coefficients

Put s= jω in equation 3, we get

$$G(j\omega)\,H(j\omega) = \frac{K\left[(1+j\omega T_1)(1+j\omega T_2)\ldots\right]}{(j\omega)^n\left[(1+j\omega T_a)(1+j\omega T_b)\ldots\right]\left[1+j2\xi\left(\frac{\omega}{\omega_n}\right)+\left(\frac{j\omega}{\omega_n}\right)^2\right]} \qquad (4)$$

$$= \frac{K\left[(1+j\omega T_1)(1+j\omega T_2)\ldots\right]}{(j\omega)^n\left[(1+j\omega T_a)(1+j\omega T_b)\ldots\right]\left[1+j2\xi u-u^2\right]} \qquad (5)$$

Where $u = \omega/\omega_n$

From equation 5, the magnitude of G(jω)H(jω) in decibels is given by

From equation (11.9), the magnitude of G (jω) H (jω) in decibels is given by

$$20\log_{10}|G(j\omega)H(j\omega)| = 20\log_{10}K + (20\log_{10}|1+j\omega T_1| + 20\log_{10}|1+j\omega T_2| \ldots)$$
$$-[20n\log\omega + 20\log|1+j\omega T_a| + 20\log|1+j\omega T_b| + \ldots] - 20\log_{10}|(1-u^2)+j2\xi u| \qquad (6)$$

or
$$20\log_{10}|G(j\omega)H(j\omega)| = [20\log K + 20\log\sqrt{1+\omega^2T_1^2} + 20\log\sqrt{1+\omega^2T_2^2} + \ldots]$$
$$-20N\log\omega - 20\log\sqrt{1+\omega^2T_a^2} - 20\log\sqrt{1+\omega^2T_b^2}\ \ldots$$
$$\ldots - 20\log\sqrt{(1-u^2)^2+4\xi^2u^2} \qquad (7)$$

and phase angle of G (jω) H (jω) is given by

$$\angle G(j\omega)H(j\omega) = \angle K + \angle(1+j\omega T_1) + \angle(1+j\omega T_2) + \ldots$$
$$-\angle(j\omega)^n - \angle(1+j\omega T_a) - \angle(1+j\omega T_b)\ldots$$
$$-\angle(1-u^2+j2\xi u) \qquad (8)$$

$$\angle G(j\omega)H(j\omega) = 0^\circ + \tan^{-1}\omega T_1 + \tan^{-1}\omega T_2 + \ldots$$
$$-(90\times n) - \tan^{-1}\omega T_a - \tan^{-1}\omega T_b\ldots$$
$$-\tan^{-1}\left(\frac{2\xi u}{1-u^2}\right) \qquad (9)$$

Generally there are the following seven simple types of factors in G (jω) H (jω) :

(i) Constant K

(ii) Zeros at origin $(j\omega)^{+n}$

(iii) Poles at the origin $(j\omega)^{-n}$

(iv) Simple zero on real axis $(1 + j\omega T)$

(v) Simple pole on real axis $\frac{1}{(1+j\omega T)}$

(vi) Complex conjugate pole $\frac{1}{(1+j2\xi u-u^2)}$

(vii) Complex conjugate zero $(1 + j2\xi u - u^2)$

Procedure for plotting Bode plot:

Step 1: Rewrite the open loop transfer function in the time constant form as given in equation 4.

Step 2: Identify the corner frequencies associated with each factor of the transfer function.

Step 3: After knowing the corner frequencies, draw the asymptotic magnitude plot. This plot consists of straight line segments with line slope changing at each corner frequency as follows.

(i) + 20 db / decade for a zero and + 20n db/decade for a zero of multiplicity n.

(ii) -20db/decade for a pole and - 20n db/decade for a pole of multiplicity n.

(iii) + 40db/decade for a complex conjugate zero and + 40n db/decade for a complex Conjugate zero of multiplicity n.

(iv)-40db/decade for a complex conjugate pole and - 40n db/decade for a complex Conjugate pole of multiplicity n.

Step 3: Initial slope of Bode plot are calculated as follows.

(i) For type zero system draw a line up to first (lowest) corner frequency having 0 db/decade slope.

(ii) For type one system dra w a line having slope of -20db/decade up to ω=K. Mark first (lowest) corner frequency.

(ii) For type two system draw a line having slope of -40db/decade up to ω=√K and so on. Mark first (lowest) corner frequency.

Step 4: Draw a line up to second corner frequency by adding the slope of next pole or zero to the previous slope and so on.

Step 5: Calculate phase angle for different values of ω from the equation 9 and join all points.

Note − The corner frequency (ω=1/K) is the frequency at which there is a change in the slope of the magnitude plot.

Example 1: Draw the bode plot for unity feedback control system having G(s)= $\frac{1000}{(s+100)}$.

Solution:

Step1: Open-loop transfer function in time constant form is given by

$$G(s)H(s)=\frac{1000}{(s+100)}$$

$$=\frac{1000}{100\frac{(s+100)}{100}}=\frac{10}{(1+0.01s)} \text{ (Time constant form)}$$

Put s= jω

$$G(j\omega)H(j\omega)=\frac{10}{(1+j0.01\omega)}$$

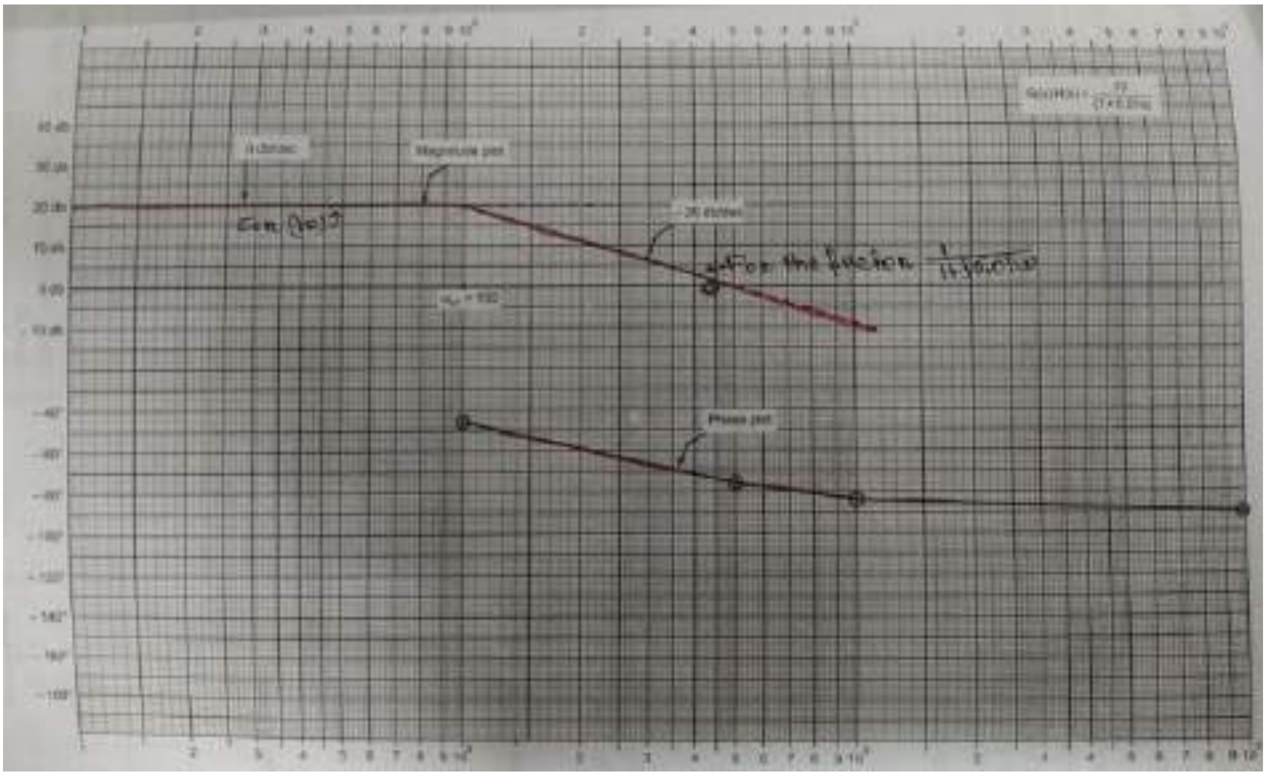
Step 2: Corner frequency ω = 1/0.001 = 100

Step 3: There is one pole on the real axis hence magnitude plot is a straight line having slope of -20 db/decade.

Step 4: As the system is type zero system so magnitude plot is a straight line parallel to o db axis and having magnitude $20\log_{10}K= 20\log_{10}10= 20$db.

Step 5: phase angle $\phi = -\tan^{-1} 0.01\omega$. The table shows value of ϕ when ω varies from 0 to ∞.

S. No.	ω	ϕ
1.	0	0°
2.	10	– 5.7°
3.	100	– 45°
4.	500	– 78.7°
5.	1000	– 84.29°
6.	10,000	– 90°



Example 2: Draw the bode plot for unity feedback control system having

G(s) = 5(s+2)/s(s+10)

Put s= jω

$$G(j\omega) = \frac{5(j\omega+2)}{j\omega(j\omega+10)} = \frac{5\times 2\left(1+\frac{j\omega}{2}\right)}{10j\omega\left(1+\frac{j\omega}{10}\right)} = \frac{1+\frac{j\omega}{2}}{j\omega\left(1+\frac{j\omega}{10}\right)}$$

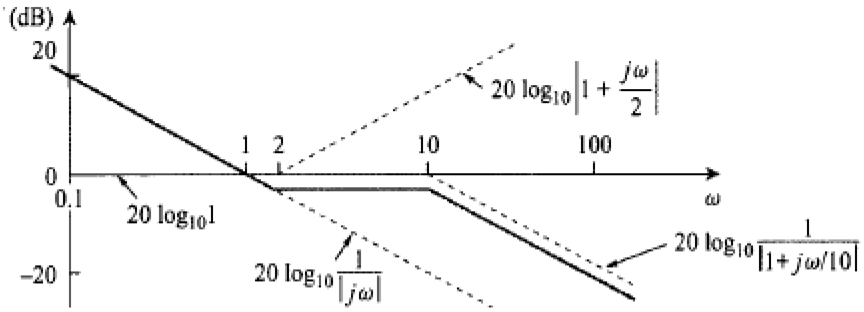
$$|G(j\omega)|\angle G(j\omega) = \frac{\left|1+\frac{j\omega}{2}\right|}{|j\omega|\left|1+\frac{j\omega}{10}\right|}\angle\left[90^\circ+\tan^{-1}\left(\frac{\omega}{2}\right)-\tan^{-1}\left(\frac{\omega}{10}\right)\right]$$

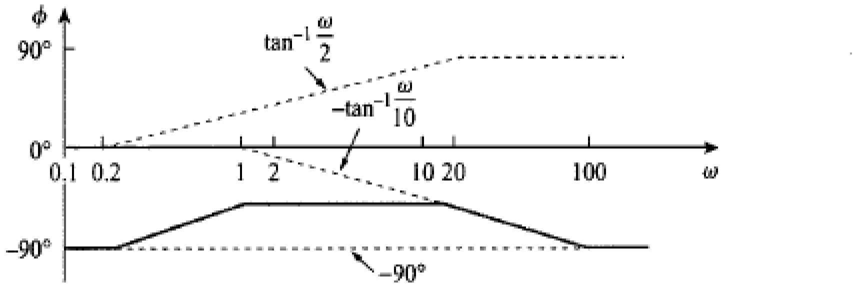
Magnitude plot:

Sl. No	Factor	Corner Frequency	Slope	Asymptotic log magnitude
1	$\frac{1}{j\omega}$	None	-20 db/decade	Straight line of slope -20 db/decade and intersecting the 0 dB axis at ω=K=1and extend upto first corner frequency 2.
2	1+0.5 $j\omega$	2	+20 db/decade	Draw a net slope (-20) + (+20)= 0 db/decade from corner frequency 2 to the next corner frequency 10.
3	$\frac{1}{1+0.1\,j\omega}$	10	-20 db/decade	Draw the net slope of 0+(-20)= -20 db/decade from corner frequency 10 to ∞.

Note: Arrange the table in increasing order of corner frequency.

For different value of ω calculate phase angle ∠G(jω) and join all the points by free hand.





Computation of Gain Margin and Phase Margin

From the Bode plots, we can say whether the control system is stable, marginally stable or unstable based on the values of these parameters.

- Gain cross over frequency and phase cross over frequency
- Gain margin and phase margin

Phase Cross over Frequency

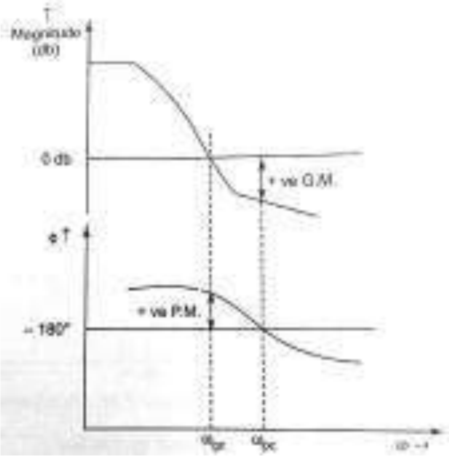
The frequency at which the phase plot is having the phase of -180^0 is known as **phase cross over frequency**. It is denoted by ω_{pc}. The unit of phase cross over frequency is **rad/sec**.

Gain Cross over Frequency

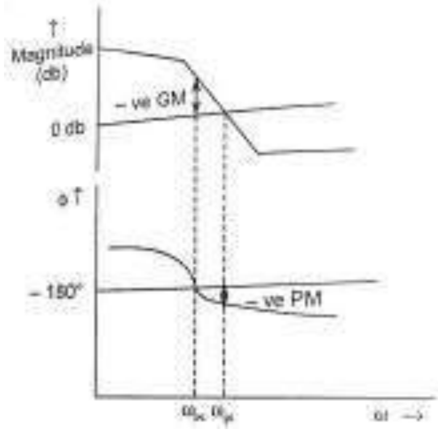
The frequency at which the magnitude plot is having the magnitude of zero dB is known as **gain cross over frequency**. It is denoted by ω_{gc}. The unit of gain cross over frequency is **rad/sec**.

The stability of the control system based on the relation between the phase cross over frequency and the gain cross over frequency is listed below.

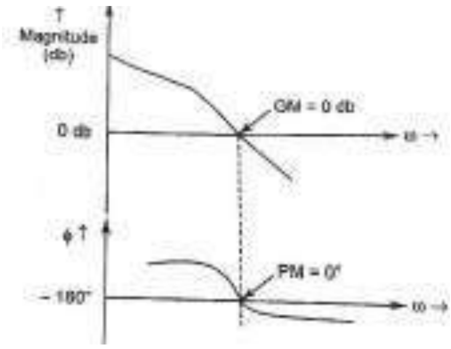
- If the phase cross over frequency ω_{pc} is greater than the gain cross over frequency ω_{gc}, then the control system is **stable**.
- If the phase cross over frequency ω_{pc} is equal to the gain cross over frequency ω_{gc}, then the control system is **marginally stable**.
- If the phase cross over frequency ω_{pc} is less than the gain cross over frequency ω_{gc}, then the control system is **unstable**.



$\omega_{pc} > \omega_{gc}$, GM & PM are +ve
Stable System



$\omega_{pc} < \omega_{gc}$, GM & PM are –ve
Un-stable System



$\omega_{pc} = \omega_{gc}$, GM= PM=0
marginally stable system

Gain Margin

Gain margin GM is defined as the margin in gain allowable by which gain can be increased till system reaches on the verge of instability. It is equal to negative of the magnitude in dB at phase cross over frequency. Mathematically

$$GM=20\log_{10}\left(\frac{1}{|G(j\omega)|_{\omega=\omega_{pc}}}\right)= -\,20\log_{10}|G(j\omega)|_{\omega=\omega_{pc}}$$

The unit of gain margin (GM) is dB.

Phase Margin

Phase margin can be defined as the amount of additional phase lag which can be introduced in the system till the system reaches on the verge of instability. The formula for phase margin PM is

$$PM=[\angle G(j\omega)|_{\omega=\omega_{gc}}] - (-180^{o})$$

$$= 180^{\circ} + [\angle(j\omega)|_{\omega=\omega_{gc}}]$$

The unit of phase margin is degrees.

The stability of the control system based on the relation between gain margin and phase margin is listed below.

- If both the gain margin GM and the phase margin PM are positive, then the control system is **stable**.
- If both the gain margin GM and the phase margin PM are equal to zero, then the control system is **marginally stable**.
- If the gain margin GM and / or the phase margin PM are/is negative, then the control system is **unstable**.

Example3: A unity feedback control system has

$$G(s)=\frac{20}{s(1+0.1s)(1+0.01s)}$$

Draw the bode plot. Find Gain crossover frequency, phase crossover frequency, gain margin and phase margin.

Solution: Put s= jω in open loop transfer function

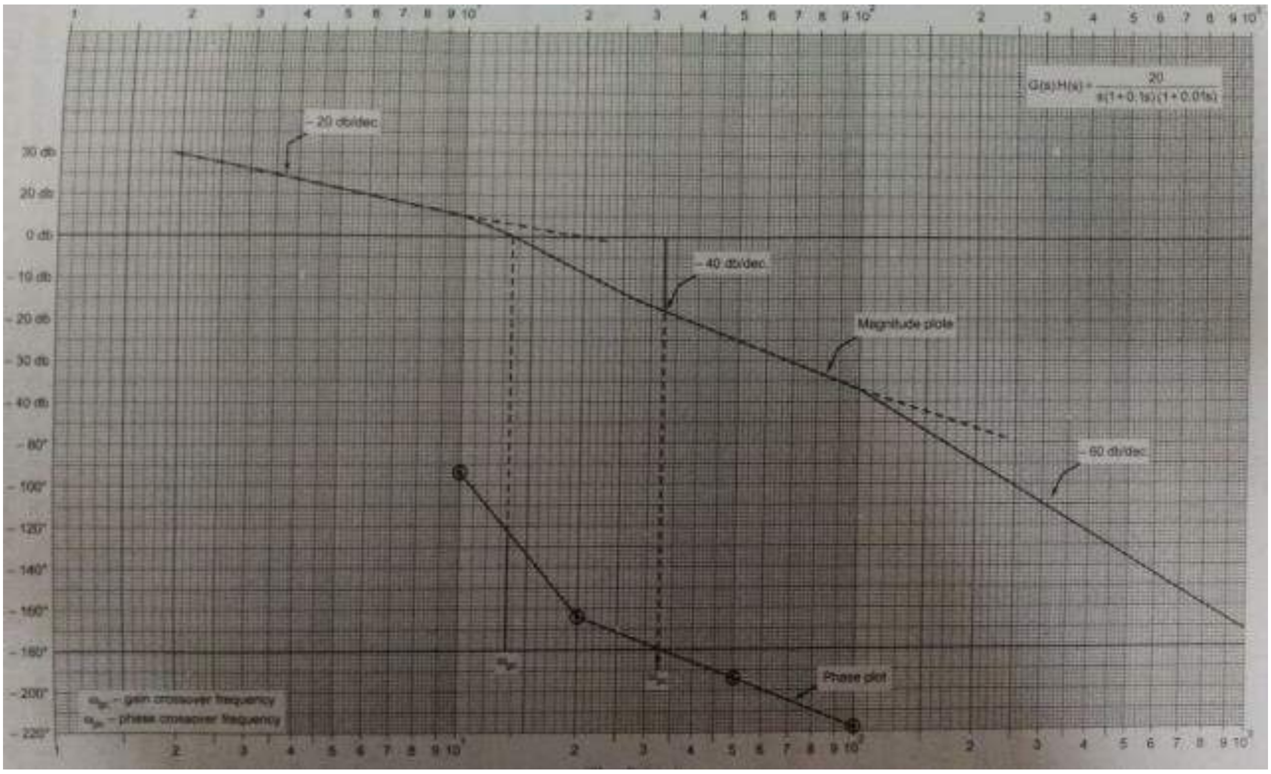
$$G(s)=\frac{20}{j\omega(1+0.1j\omega)(1+0.01j\omega)}$$

$$|G(j\omega)|\angle G(j\omega)=\frac{20}{-\omega^2\sqrt{1+(0.1\omega)^2}\sqrt{1+(0.01\omega)^2}}\angle-90^\circ-tan^{-1}0.1\omega-tan^{-1}0.01\omega$$

Sl. No	Factor	Corner Frequency	Slope	Asymptotic log magnitude
1	$\frac{20}{j\omega}$	None	-20 db/decade	Straight line of slope -20 db/decade and intersecting the 0 dB axis at ω=K=20 and extend upto first corner frequency 10.
2	$\frac{1}{1+0.1\,j\omega}$	10	20 db/decade	Draw a net slope (-20) + (-20)= -40 db/decade from corner frequency 10 to the next corner frequency 100.
3	$\frac{1}{1+0.01\,j\omega}$	100	-20 db/decade	Draw the net slope of (- 40)+(-20)= -60 db/decade from corner frequency 100 to ∞.

The table shown below shows phase angle for the different value of ω.

S.No.	ω	ϕ
1.	0	– 90°
2.	10	– 95.71°
3.	20	– 164.74°
4.	50	– 195.25°
5.	100	– 219.3°



From the plots

1. Gain crossover frequency $\boldsymbol{\omega_{gc}}$ = 13.5
2. Phase crossover frequency $\boldsymbol{\omega_{pc}}$ = 33
3. $\omega_{pc} > \omega_{gc}$, GM & PM are +ve, hence the system become stable.
4. Gain Margin = + 15 db
5. Phase Margin= 180^o-($+124^o$)= $+56^o$

All pass and minimum phase system

If all the poles and zeros of any transfer function lie o the left half of s-plane, such type of transfer function is known as ***minimum phase transfer function.***

The transfer function having a pole-zero pattern which is antisymmetric about the imaginary axis i.e for every pole in the left half plane, there is a zero in the mirror image position. This type of transfer function is known as ***all pass transfer function.***

A common example of such transfer function is

$$G(j\omega)=\frac{1-j\omega T}{1+j\omega T} \qquad (1)$$

Pole zero configuration of equation 1is shown below:

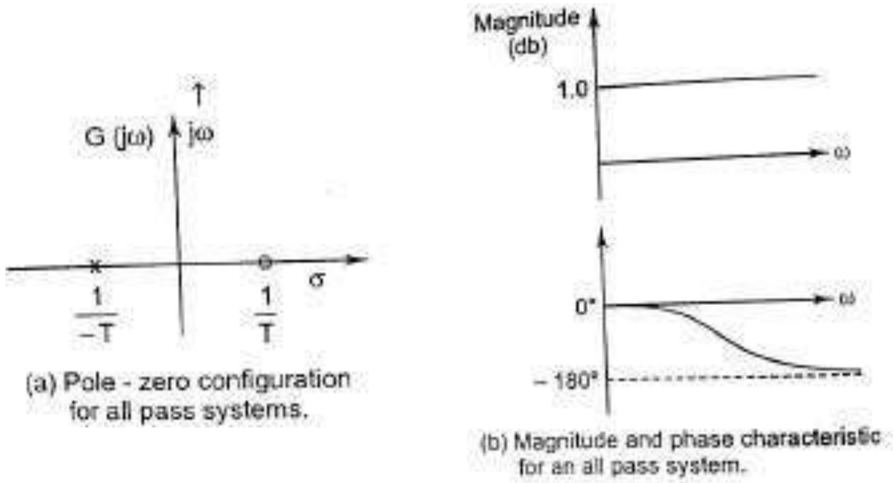


Figure 1. All pass system

All pass transfer function has a magnitude of unity at all frequency and a phase angle of (- 2 $\tan^{-1}\omega T$)which varies from 0° to -180° as ω increases from 0 to ∞. *The property of unit magnitude at all frequencies applies to all transfer function with antisymmetric pole-zero pattern.* Physical systems with this property are called all-pass system.

Now consider the case where the transfer function has poles in the left half s-plane and zero in both left and right half s-plane. Poles are not permitted to lie in the right half s-plane because such a system would be unstable. Consider the following transfer function

$$G_1(j\omega)= \frac{1-j\omega T}{(1+j\omega T_1)(1+j\omega T_2)} \tag{2}$$

Whose pole zero pattern is shown in figure . This transfer function may be rewritten as

$$G_1(j\omega)= \left[\frac{1+j\omega T}{(1+j\omega T_1)(1+j\omega T_2)}\right] \left[\frac{1-j\omega T}{1+j\omega T}\right] = G_2(j\omega)\, G(j\omega) \tag{3}$$

Which is now become the product of two transfer function $G_2(j\omega)$ i.e minimum phase transfer function shown in figure (2b) and $G(j\omega)$ i.e all pass transfer function shown in figure (2c). It is clear that $G_1(j\omega)$ and $G_2(j\omega)$ have identical curve of magnitude Vs frequency but their phase Vs frequency curve are different as shown in figure(3). $G_2(j\omega)$ having a smaller range of phase angle than $G_1(j\omega)$. A transfer function which has one or more zeros and no pole in the right half s-plane is known as non- minimum phase transfer function.

In general if the transfer function has any zeros in the right half s-plane, it is possible to extract them one by one by associating them with all-pass transfer function as shown in figure(2a).

A common example of a non-minimum phase element is transportation lag which has transfer function

$$G(j\omega)= e^{-j\omega T}= 1\angle -\omega T \text{ rad} = 1\angle -57.3\omega T \text{ degree}$$

Other possible non-minimum transfer function are:

1. where more than one possible signal paths are available between input and output as in lattice network.
2. When there is inductive coupling between input and output in addition to conduction.

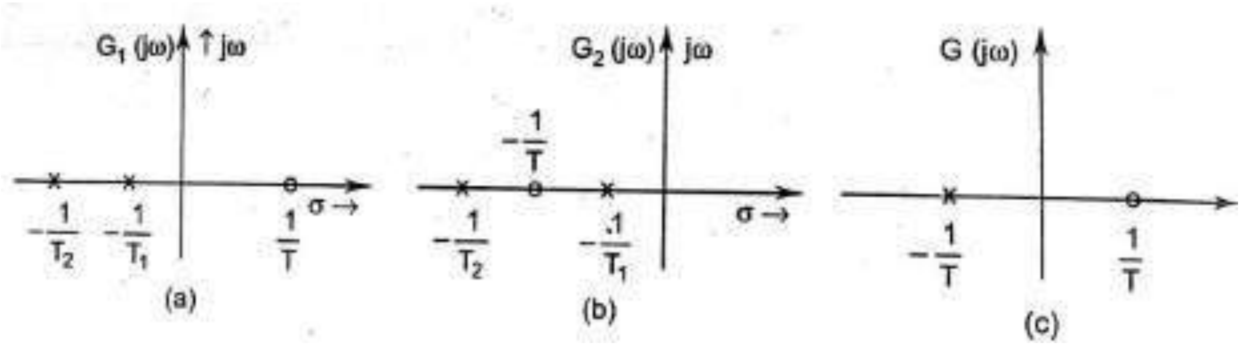


Figure 2

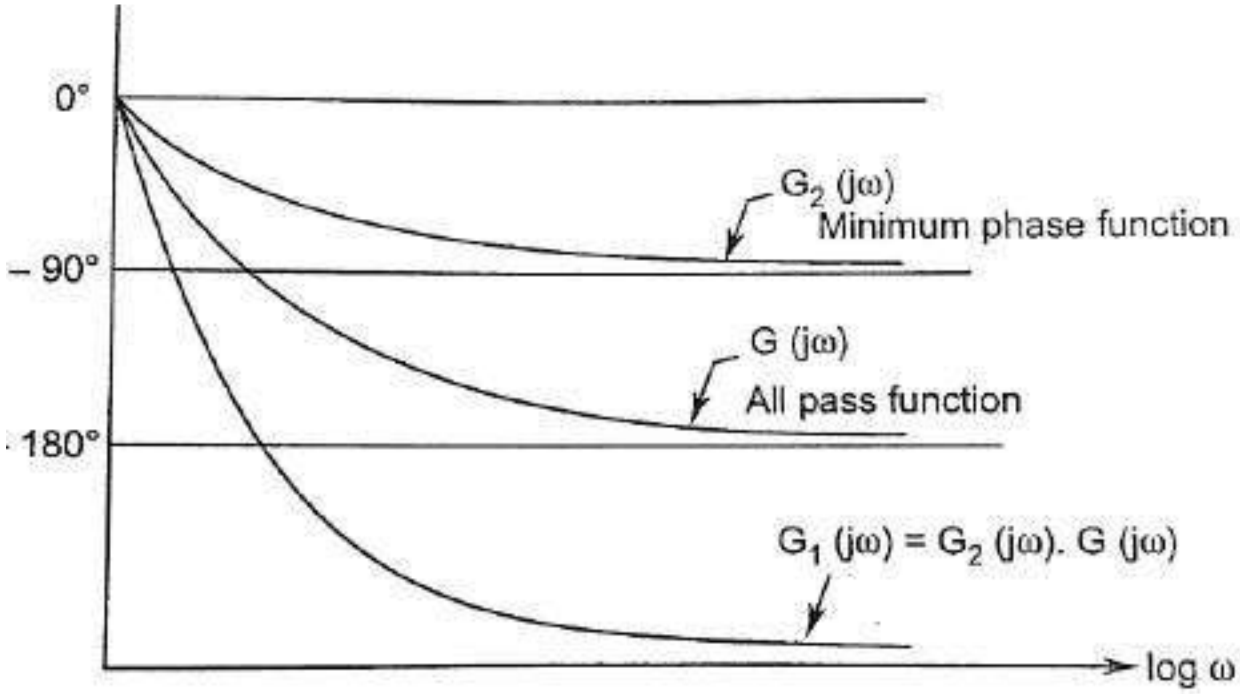


Figure 3 Phase Vs frequency graph

Polar Plots

Polar plot is a plot which can be drawn between magnitude and phase. It is a plot of magnitude $|(j\omega)|$ versus phase angle$\angle G(j\omega)$on polar co-ordinates as input frequency (ω) is varied from 0 to ∞. Here, the magnitudes are represented by normal values only.

The polar form of G(jω) is

$$G(j\omega) = |G(j\omega)| \angle G(j\omega)$$

Rules for Drawing Polar Plots

Follow these rules for plotting the polar plots.

Step1. Substitute, s=jω in the open loop transfer function.
Step2. Write the expressions for magnitude and the phase of G(jω)

Step3. Find the starting magnitude and the phase of G(jω) by substituting ω=0. So, the polar plot starts with this magnitude and the phase angle.
Step4. Find the ending magnitude and the phase of G(jω) by substituting ω=∞. So, the polar plot ends with this magnitude and the phase angle.
Step5. Check whether the polar plot intersects the real axis, by making the imaginary term of G(jω) equal to zero and find the value(s) of ω.
Step6. Determine the intersection of polar plot with real axis and imaginary axis, as follows:

i. Rationalise the function G(jω) and separate the real and imaginary parts.

ii. Intersection with imaginary axis: equate the real term of |G(jω)| to zero and find the value of frequency (ω) at which the polar plot intersects the imaginary axis. Now put this value of ω into |G(jω)|. Which gives |G(jω)| at this point of intersection.

iii. Intersection with real axis: equate the imaginary term of |G(jω)| to zero and find the value of frequency (ω) at which the polar plot intersects the real axis. Now put this value of ω into |G(jω)|. Which gives |G(jω)| at this point of intersection.

Step7. By using this information, plot the points on the complex plane. Make the arrow on the plot for increasing frequency from 0 to ∞.

Example1: Consider the open loop transfer function of a closed loop control system.

$$G(s)=\frac{1}{(1+sT_1)(1+sT_2)}$$

Draw the polar plot.

Step 1 − Substitute, s=jω in the open loop transfer function.

$$G(j\omega)=\frac{1}{(1+j\omega T_1)(1+j\omega T_2)}$$

The magnitude of the open loop transfer function is

$$|G(j\omega)|=\frac{1}{\sqrt{1+(\omega T_1)^2}\sqrt{1+(\omega T_2)^2}}$$

The phase angle of the open loop transfer function is

$$\angle G(j\omega)=-\tan^{-1}\omega T_1-\tan^{-1}\omega T_2$$

Step 2 − The following table shows the magnitude and the phase angle of the open loop transfer function at ω=0 rad/sec and ω=∞ rad/sec.

Frequency (rad/sec)	Magnitude	Phase angle(degrees)
0	1	0
∞	0	-180°

So, the polar plot starts at $(1,0^0)$ and ends at $(0,-180^0)$. The first and the second terms within the brackets indicate the magnitude and phase angle respectively.

Step 3 − This polar plot will intersect the negative imaginary axis. The phase angle corresponding to the negative imaginary axis is -90^0 or 270^0. So, by equating the phase angle of the open loop transfer function to either -90^0 or 270^0, we will get the ω value as

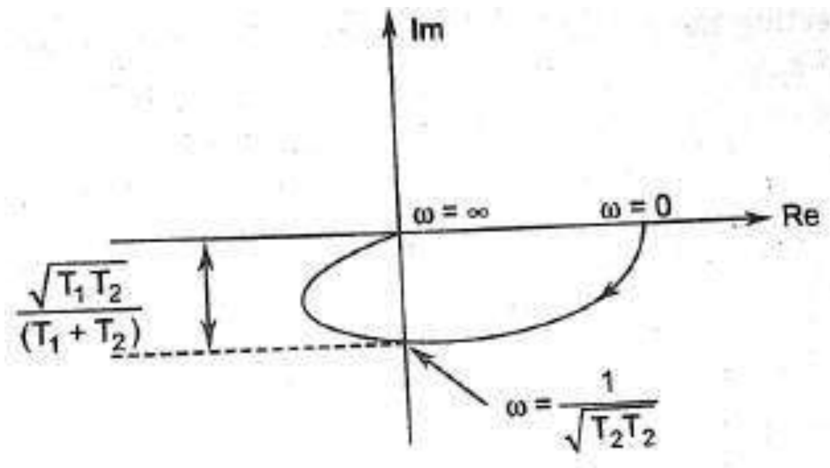
$$\angle G(j\omega) = -\tan^{-1}\omega T_1 - \tan^{-1}\omega T_2 = -90^0$$

$$\Rightarrow \frac{T_1 + \omega T_2}{1-\omega^2 T_1 T_2} = \infty \Rightarrow \omega = \frac{1}{\sqrt{T_1 T_2}}$$

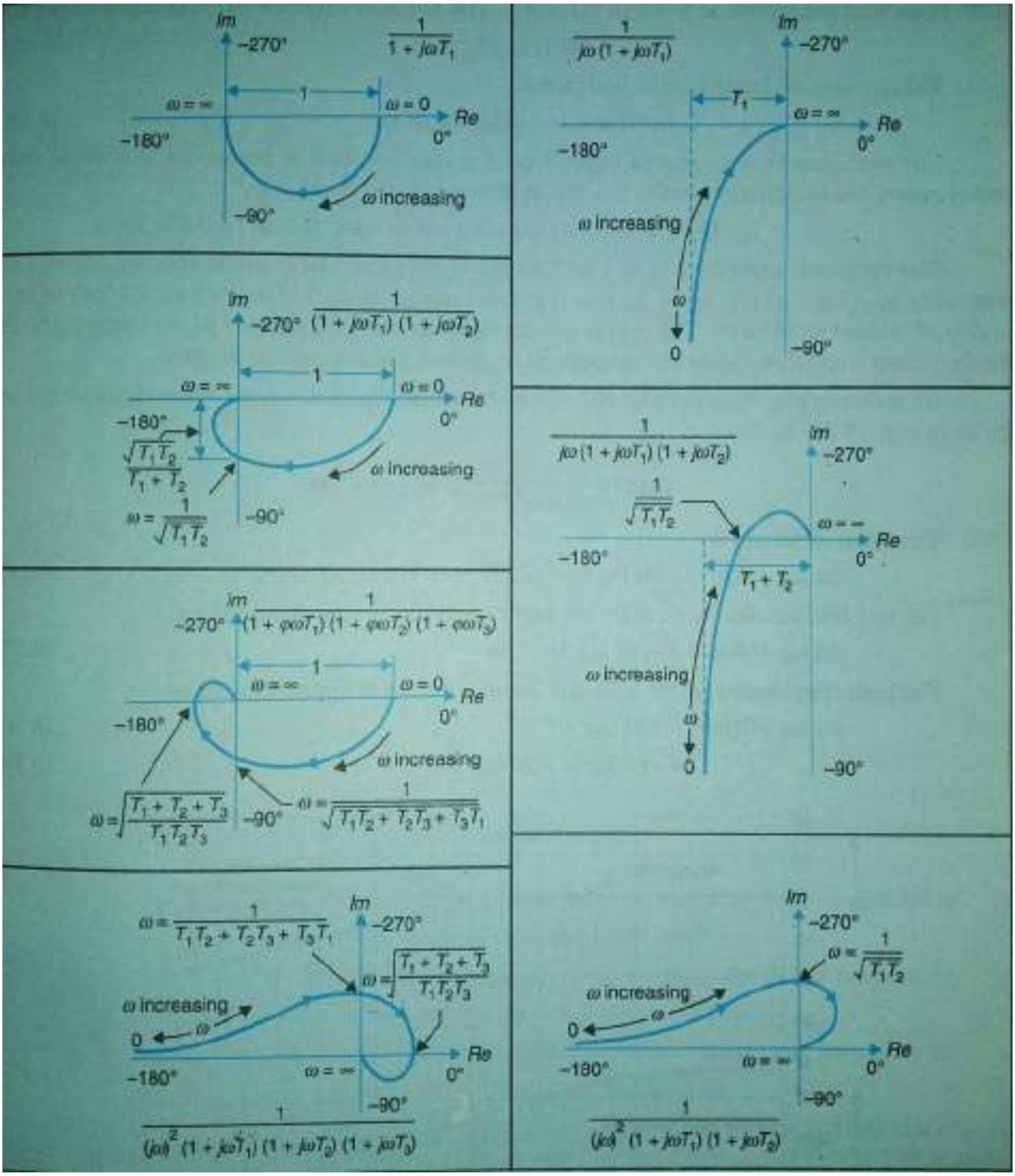
By substituting $\omega = \frac{1}{\sqrt{T_1 T_2}}$ in the magnitude of the open loop transfer function, we will get

$$|G(j\omega)| = \frac{1}{\sqrt{1+\left(\frac{1}{\sqrt{T_1T_2}}T_1\right)^2}\sqrt{1+\left(\frac{1}{\sqrt{T_1T_2}}T_2\right)^2}} = \frac{\sqrt{T_1 T_2}}{T_1 + T_2}$$

So, we can draw the polar plot with the above information on the polar graph sheet.



The following table shows polat plot for different type of control system:



CHAPTER- V
Nyquist Plots

Introduction:

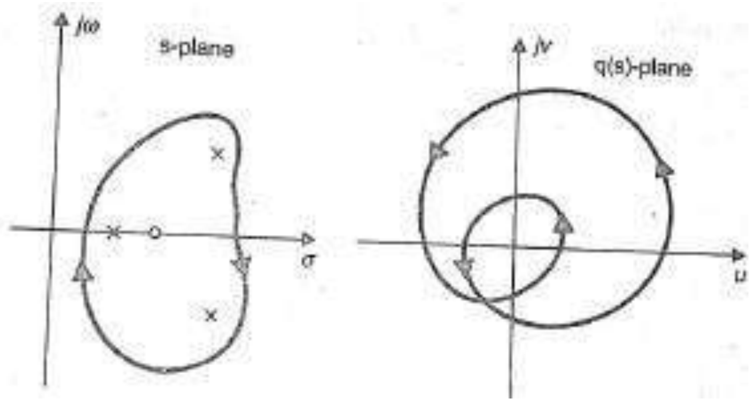
Nyquist plots are the continuation of polar plots for finding the stability of the closed loop control systems by varying ω from −∞ to ∞. That means, Nyquist plots are used to draw the complete frequency response of the open loop transfer function.

Principle of argument

The Nyquist stability criterion works on the **principle of argument**. It states that if there are P poles and Z zeros are enclosed by the 's' plane contour, then the corresponding G(s)H(s) plane must encircle the origin P−Z times. So, we can write the number of encirclements N as,
N=P−Z

- If the 's' plane contour contains only poles, then the direction of the encirclement in the q(s) plane will be opposite (counter clock wise) to the direction of 's' plane contour.
- If the 's' plane contour contains only zeros, then the direction of the encirclement in the q(s) plane will be in the same (clock wise) direction as that of 's' plane contour.

For example, in case of 1 zero and 3 poles enclosed by the s- plane contour, the net encirclement of the origin by the q(s) plane contour is (3-1) two counter-clockwise revolution as shown in figure below. This relationship between the enclosure of poles and zeros of G(s)H(s) b the s-plane contour and the encirclement of the origin by G(s)H(s) contour is commonly known as principle of argument.



Nyquist stability criterion

The characteristics equation of a system is
q(s) = 1+G(s)H(s)
The standard pole zero form of the OLTF G(s)H(s) is

$$G(s)H(s) = K\frac{(s+z_1)(s+z_2)--------(s+z_m)}{(s+P_1)(s+P_2)--------(s+P_n)} \quad (1)$$

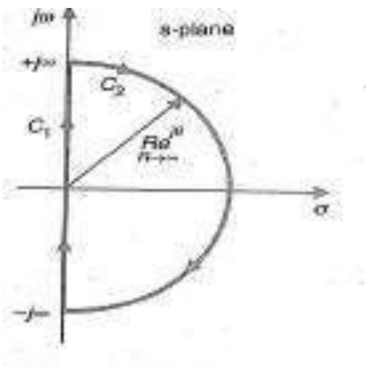
$\therefore$ q(s) = 1+ $K\frac{(s+z_1)(s+z_2)--------(s+z_m)}{(s+P_1)(s+P_2)---------(s+P_n)}$

$$= \frac{(s+P_1)(s+P_2)---------(s+P_n)+K(s+z_1)(s+z_2)--------(s+z_m)}{(s+P_1)(s+P_2)---------(s+P_n)}$$

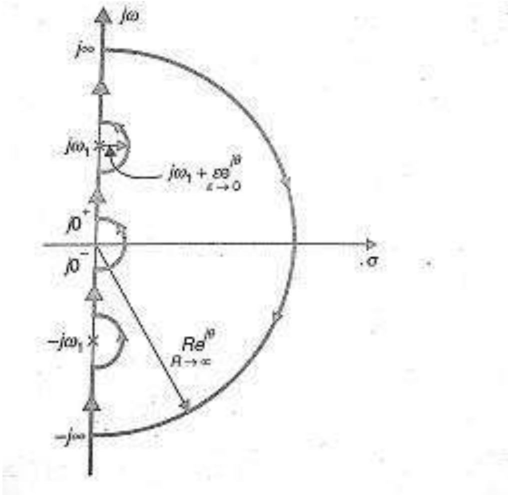
$$= \frac{(s+z^F{}_1)(s+z^F{}_2)---------(s+z^F{}_n)}{(s+P_1)(s+P_2)---------(s+P_n)} \quad (2)$$

From the above equation it is seen that the zeros of q(s) are the root of the characteristics equation and the poles q(s) are same as the poles of open loop system. For the system to be stable, the roots of the characteristics equation and hence the zeros of q(s) must lie in the left half s-plane. It is important to note that even if some of the open-loop poles lie in the right half s-plane all the zeros of q(s) i.e, the closed-loop poles may lie in the left half s-plane. It means that an open-loop unstable system may lead to a closed-loop stable system.

In order to investigate the presence of any zero of q(s) in the right half of s-plane, a contour to be chosen which completely encloses the right half of s-plane called as **Nyquist contour**. It is directed clockwise and consist of an infinite line segment C_1 and an arc C_2 of infinite redius.



The Nyquist contour.



Indented Nyquist contour for jω-axis open-loop poles.

As the Nyquist contour encloses all the right half s-plane poles and zeros of q(s), let there are 'z' zeros and 'P' poles in the right half of s-plane. As s moves along the nyquist contour in the s-plane, a closed contour Γ_q is traversed in q(s) plane which encloses the origin N (=P-Z) times in anticlockwise direction.

For the system to be stable, there should be no zeros of q(s) in the right half of s-plane i.e,

$$Z = 0$$

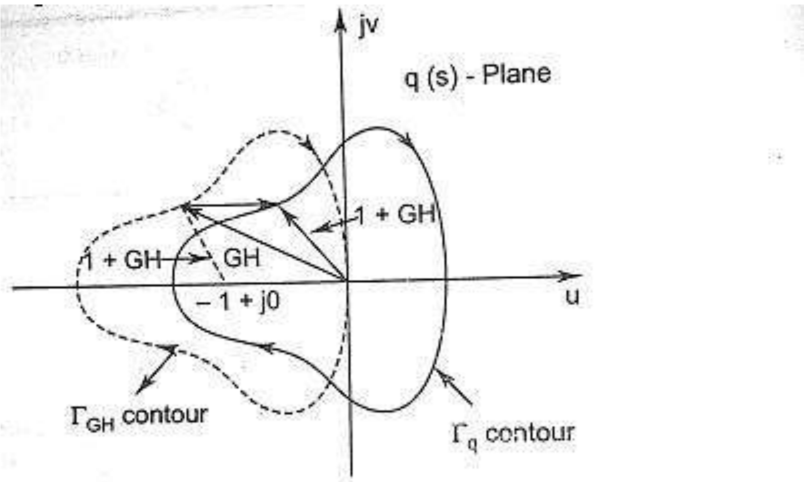
$$\text{So } N = P$$

The above equation implies that for a close loop system to be stable, the number of counter-clockwise encirclement of the origin of the q(s) plane by the contour Γ_q should be equal the number of the right half s-plane poles of q(s) which are also the poles of open-loop transfer function G(s)H(s).

The open-loop transfer function can be written as

$$G(s)H(s)=q(s)- 1= [1+ G(s)H(s)]-1 \quad (3)$$

Therefore the contour Γ_{GH} of G(s)H(s) corresponding to the nyquist contour in the s-plane is the same as contour Γ_q of q(s) (=1+ G(s)H(s)) drawn from the point (-1+j0). Thus the encirclement of the origin by the contour Γ_q of q(s) is equivalent to the encirclement of the point (-1+j0) by the contour Γ_{GH} of G(s)H(s) as shown below.



Contour Γ_{GH} of G(s) H (s) corresponding to Nyquist contour.

Statement of nyquist stability criterion:

1. If the contour Γ_{GH} corresponding to the Nyquist contour in the s-plane encircles the point (-1+j0) in the counter-clockwise direction as many times as the number of right half s-plane pole of G(s)H(s), the close loop system is stable.
2. The closed loop system is stable if the contour Γ_{GH} does not encircles the point (-1+j0).

Mapping of Nyquist contour into the contour Γ_{GH} of G(s)H(s):

1. For imaginary axis: Put s= jω in G(s)H(s) where s varies from –j∞ to +j∞ .

2. For infinite semi circle: put s= $Re^{j\theta}$ where R→∞ and θ varies from + 90° to -90°.

3. For presence of pole at origin: put s= $\in e^{j\theta}$ where ∈→∞ and θ varies from - 90° to +90°.

4. For presence of pole at imaginary axis: put s= $j\omega_1 + \in e^{j\theta}$ where ∈→∞ and θ varies from - 90° to +90°.

Hence the complete contour Γ_{GH} is the polar plot of G(jω)H(jω) with varies from ω -∞ to +∞.

Nyquist stability criterion applied to inverse Polar plot:

It is more convenient to work with inverse function 1/ G(jω)H(jω) rather than the direct function G(jω)H(jω). Here we will see that the Nyquist stability criterion for direct polar plot can be extended for use to inverse polar plot after minor modification.

Let us consider a open-loop transfer function:

$$G(s)H(s) = K\frac{(s+z_1)(s+z_2)-------(s+z_m)}{(s+P_1)(s+P_2)--------(s+P_n)} \qquad (4)$$

For the system to be stable none of the roots of the characteristics equation should lie in the right half s-plane or on the jω-axis. The characteristics equation is

$$q(s)= 1+ G(s)H(s) = \frac{(s+z^F_1)(s+z^F_2)---------(s+z^F_n)}{(s+P_1)(s+P_2)--------(s+P_n)} \quad (5)$$

Dividing equation 5 by 4, we get

$$q'(s)= \frac{1}{G(s)H(s)} +1= \frac{(s+z^F_1)(s+z^F_2)---------(s+z^F_n)}{(s+z_1)(s+z_2)-------(s+z_m)} \quad (6)$$

From equation 5 and 6 it is seen that the zeros of q'(s) is same as the q(s), which are the roots of the characteristics equation. It is further noticed that the poles of q(s) are same as the poles of G(s)H(s), while the poles of q'(s) are same as the poles of $\frac{1}{G(s)H(s)}$ or the zeros of G(s)H(s).

It can be concluded that if $\frac{1}{G(s)H(s)}$ has P right half s-plane poles and the characteristics equation has Z right half s-plane zeros, the locus of $\frac{1}{G(s)H(s)}$ encircle the point (-1+j0) N times in counter-clockwise direction where N= P-Z.

Since for system stability no zeros of the characteristics equation locate on right half s-plane i.e , Z=0, the Nyquist stability criterion for inverse polar plots can be stated below:

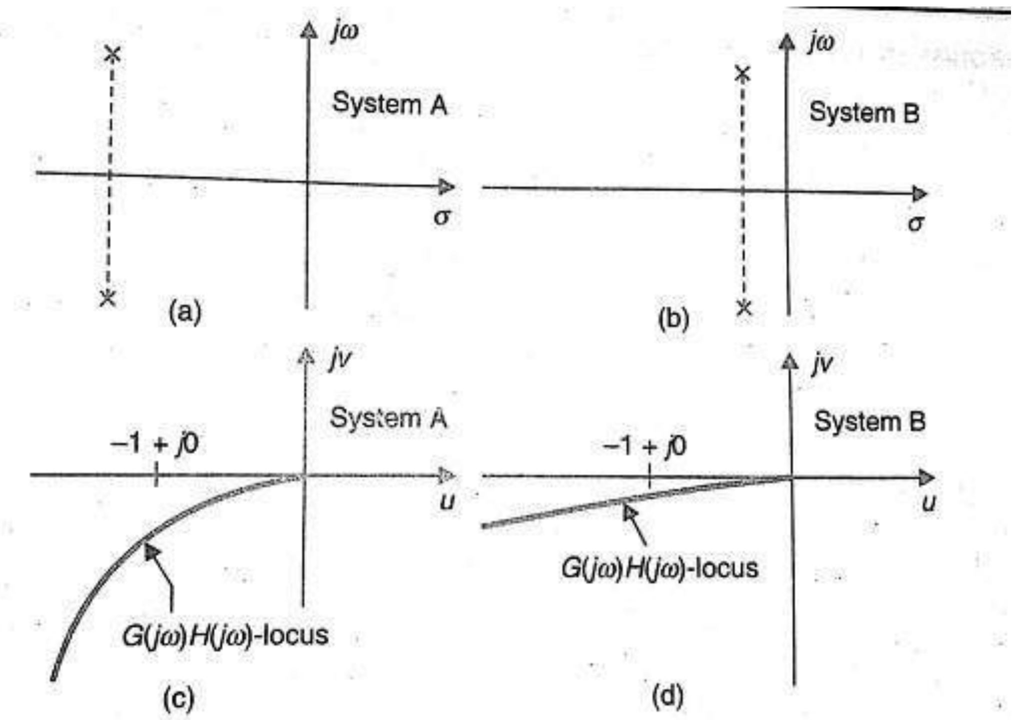
"It the Nyquist plot of $\frac{1}{G(s)H(s)}$ corresponding to the Nyquist contour in the s-plane, encircles counter-clockwise the point (-1+j0) as many times as are the number of right half s-plane pole of $\frac{1}{G(s)H(s)}$, the closed-loop system is stable. "

In special case where $\frac{1}{G(s)H(s)}$ has no pole in the right half s-plane, the close loop system is stable provided the net encirclement of (-1+j0) point by the Nyquist plot of $\frac{1}{G(s)H(s)}$ is zero.

Assessment of relative stability using Nyquist criterion:

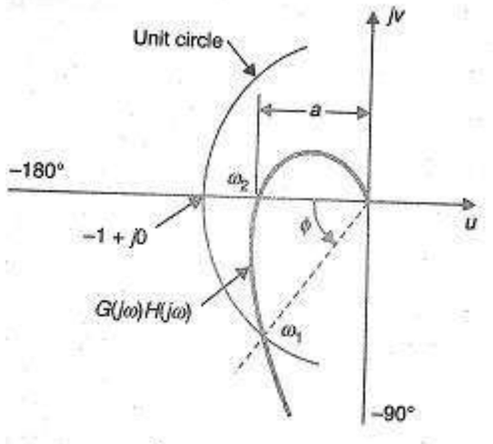
The measure of relative stability of a closed-loop systems which are open-loop stable can be analysed through the study of Nyquist plots. The stability of such system can be determined by polar plot of G(s)H(s). It can be imagined that as the polar plot gets closer to (-1+j0) point, the system tends towards instability.

Consider two different systems whose closed loop poles are shown on the s-plane in figure a and b respectively. It is seen that system A is more stable than system B because its closed-loop poles are located comparatively away to the left from jω-axis. The open-loop frequency response (polar) plots for system A and B are shown in figure 'c' and 'd', respectively. The comparison of the closed-loop pole location of these two system with their corresponding polar plot shows that as a polar plot moves closer to (-1+j0) point, the system closed-loop poles move closer to the jω-axis and hence the system becomes relatively less stable and vice versa.



Correlation between the closed-loop s-plane root locations and open-loop frequency response curves.

The figure as given below shows a G(jω)H(jω) locus which crosses the negative real axis at a frequency $\omega=\omega_2$ with an intercept of a. Let a unit circle centred at origin (passes through point -1+j0) intersect the G(jω)H(jω) locus at a frequency $\omega=\omega_1$ and let the phasor $G(j\omega_1)H(j\omega_1)$ makes an angle of ϕ with the negative real axis measured positively in counter-clockwise direction. It is observed that as G(jω)H(jω) locus approaches (-1+j0) point, the relative stability reduces.



A typical G(jω)H(jω)-locus.

Constant Magnitude Loci or Constant M Circle

The closed loop transfer function of a unity feedback system is given by

$$T(s) = \frac{C(s)}{R(s)} = \frac{G(s)}{1+G(s)}$$

or

$$T(j\omega) = \frac{C(j\omega)}{R(j\omega)} = \frac{G(j\omega)}{1+G(j\omega)}$$

Let

$$G(j\omega) = x + jy$$

$\therefore$

$$T(j\omega) = \frac{(x+jy)}{1+(x+jy)} = \frac{x+jy}{1+x+jy} \qquad (1)$$

Let magnitude of T ($j\omega$) is M, we can write

$$M = \frac{|x+jy|}{|1+x+jy|} = \frac{\sqrt{x^2+y^2}}{\sqrt{(1+x)^2+y^2}} \qquad (2)$$

Squaring both sides, we get

$$M^2 = \frac{x^2+y^2}{(1+x)^2+y^2} \qquad (3)$$

or $$M^2[(1+x)^2+y^2] = x^2+y^2$$

or $$M^2[1+x^2+2x+y^2] = x^2+y^2$$

or $$M^2[x^2+y^2+2x+1] - x^2 - y^2 = 0$$

or $$x^2(M^2-1) + 2xM^2 + y^2(M^2-1) + M^2 = 0$$

or $$x^2 + \frac{2x}{(M^2-1)}.M^2 + y^2 + \frac{M^2}{(M^2-1)} = 0$$

or $$x^2 - \frac{2x}{(1-M^2)}.M^2 + y^2 - \frac{M^2}{(1-M^2)} = 0$$

or $$x^2-\frac{2xM^2}{(1-M^2)}+y^2=\frac{M^2}{(1-M^2)}$$

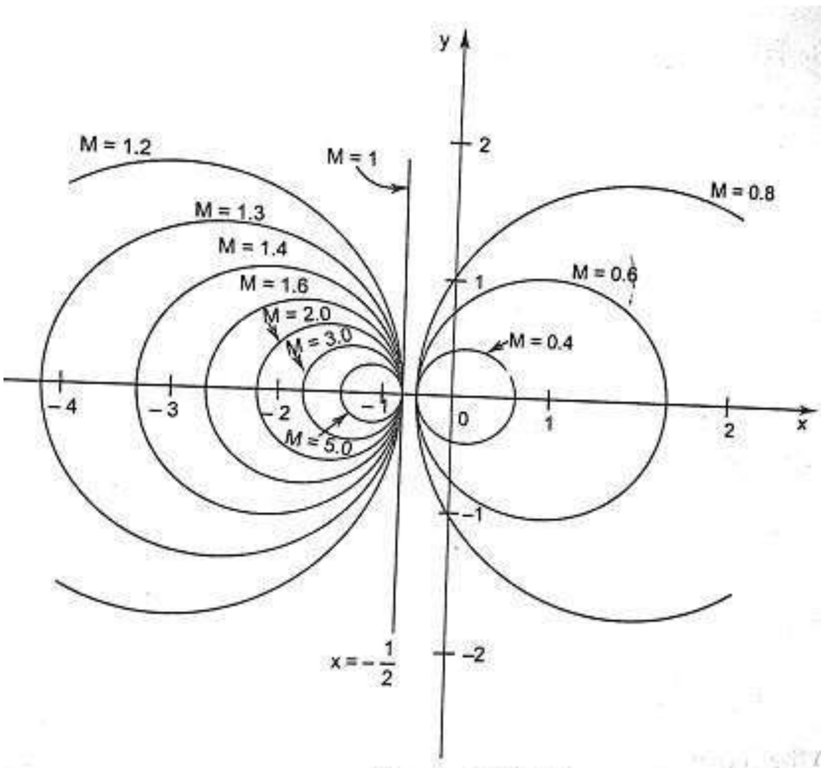
or $$x^2-\frac{2xM^2}{(1-M^2)}+\left[\frac{M^2}{(1-M^2)}\right]^2+y^2=\frac{M^2}{(1-M^2)}+\left[\frac{M^2}{(1-M^2)}\right]^2$$

or $$\left[x-\frac{M^2}{(1-M^2)}\right]^2+y^2=\frac{M^2}{(1-M^2)^2} \qquad \textbf{(4)}$$

Equation 2 represents the equation of a circle with centre at $[\frac{2}{(1-M^2)}, 0]$ having radius of $\frac{M}{(1-M^2)}$

If M=1, then Equation 3 becomes $(1+x)^2+y^2 = x^2 + y^2$ or $x=-\frac{1}{2}$ (5)

It is a equation for straight line parallel to the y-axis and passing through $(-\frac{1}{2},0)$ in the G(jω) plane. For each value of M (except M=1) we get a circle. These circles are known as Constant Magnitude Loci or Constant M Circle.



Constant M-circles.

Constant Phase Loci or Constant N Circle

From equation 1

$$T(j\omega) = \frac{C(j\omega)}{R(j\omega)} = \frac{(x+jy)}{(1+x)+jy}$$

Phase angle of T ($j\omega$) is given by

$$\angle T(j\omega) = \angle\left[\frac{C(j\omega)}{R(j\omega)}\right] = \angle\left[\frac{(x+jy)}{(1+x)+jy}\right]$$

$\Rightarrow$
$$\theta = \tan^{-1}\left(\frac{y}{x}\right) - \tan^{-1}\left(\frac{y}{1+x}\right) \qquad (6)$$

$\Rightarrow$
$$\tan\theta = \frac{\frac{y}{x} - \frac{y}{1+x}}{1 + \frac{y^2}{x(1+x)}}$$

$$= \frac{y}{x^2+x+y^2}$$

Let
$$\tan\theta = N \qquad (7)$$

$\therefore$
$$N = \frac{y}{x^2+x+y^2} \qquad (8)$$

or
$$N(x^2+x+y^2) = y$$

or
$$x^2+x+y^2-\frac{y}{N} = 0$$

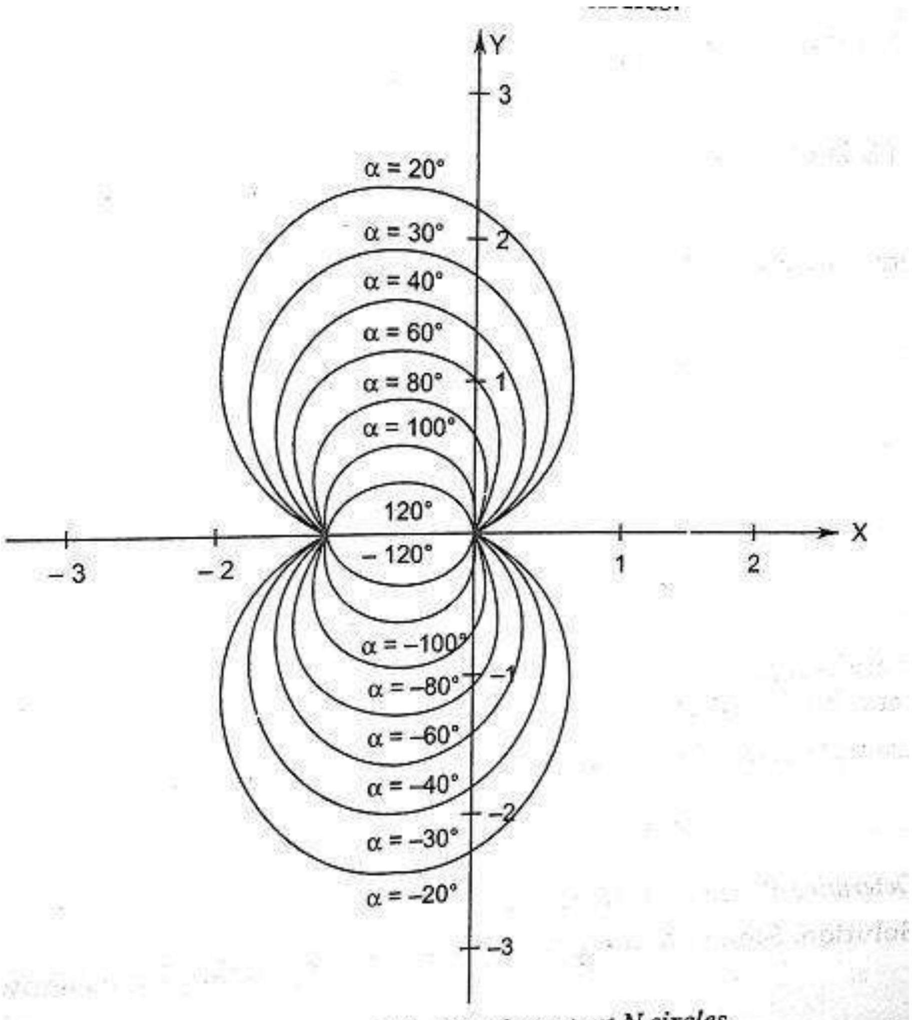
Add $\left(\frac{1}{4}+\frac{1}{4N^2}\right)$ to both sides, we get **(9)**

$$x^2+x+y^2-\frac{y}{N}+\frac{1}{4}+\frac{1}{(4N^2)} = \frac{1}{4}+\frac{1}{4N^2}$$

or
$$\left(x+\frac{1}{2}\right)^2 + \left(y-\frac{1}{2N}\right)^2 = \frac{1}{4}+\frac{1}{4N^2} \qquad (10)$$

Equation 10 represents the equation of circle with it centre at $\left(-\frac{1}{2}, \frac{1}{2N}\right)$ with radius $\sqrt{\left(\frac{1}{4}+\frac{1}{4N^2}\right)}$

For different values of N i.e, phase angle θ, equation 10 represents the family of the circles. For a particular circle, the value of N i.e, phase angle θ remain constant on it. Therefore these circle are known as constant phase loci or N circles.



A family of constant N circles.

Nichols Plot

Constant magnitude loci that are M-circles and constant phase angle loci that are N-circles are the fundamental components in designing the Nichols chart. The constant M and constant N circles in G (jω) plane can be used for the analysis and design of control systems. However the constant M and constant N circles in gain phase plane are prepared for system design and analysis as these plots supply information with fewer manipulations. Gain phase plane is the graph having gain in decibel along the ordinate (vertical axis) and phase angle along the abscissa (horizontal axis). The M and N circles of G (jω) in the gain phase plane are transformed into M and N contours in rectangular co-ordinates. A point on the constant M loci in G (jω) plane is transferred to gain phase plane by drawing the vector directed from the origin of G (jω) plane to a particular point on M circle and then measuring the length in db and angle in degree. The critical point in G (jω), plane corresponds to the point of zero decibel and -180^{o} in the gain phase plane. Plot of M and N circles in gain phase plane is known as Nichols chart /plot.

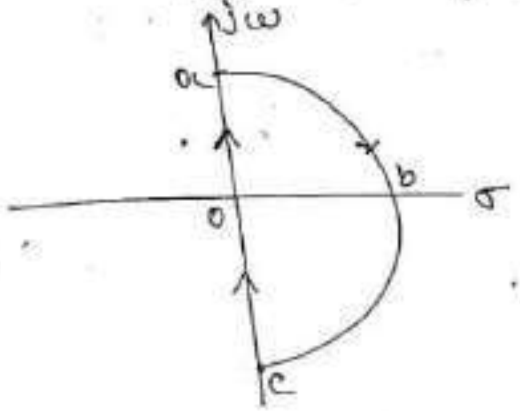
The **Nichols plot** is named after the American engineer N.B Nichols who formulated this plot. Compensators can be designed using Nichols plot. Nichols plot technique is however also used in designing of dc motor. This is used in signal processing and control design. Nyquist plot in complex plane shows how phase of transfer function and frequency variation of magnitude are related. We can find out the gain and phase for a given frequency. Angle of positive real axis determines the phase and distance from origin of complex plane determines the gain.

Example:1 - Sketch Nyquist plot for a given OLTF.

$$G(s)H(s) = \frac{K}{(T_1 s+1)(T_2 s+1)}$$

Solution

step1: First draw nyquist path for the given OLTF.



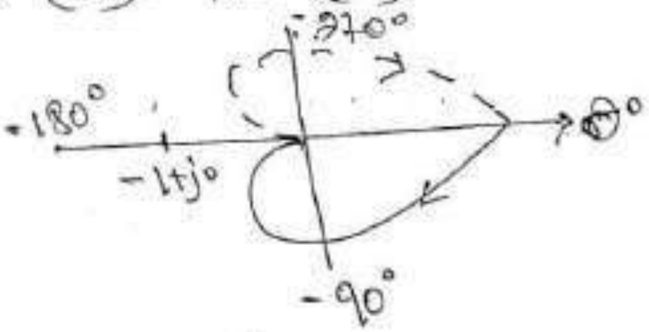
step2: For path oa put $s=j\omega$ where ω varies from 0 to ∞ & plot polar plot.

$$G(j\omega)H(j\omega) = K/(T_1 j\omega+1)(T_2 j\omega+1)$$

$$M = |G(j\omega)H(j\omega)| = K/\sqrt{\{(T_1\omega)^2+1\}\{(T_2\omega)^2+1\}}$$

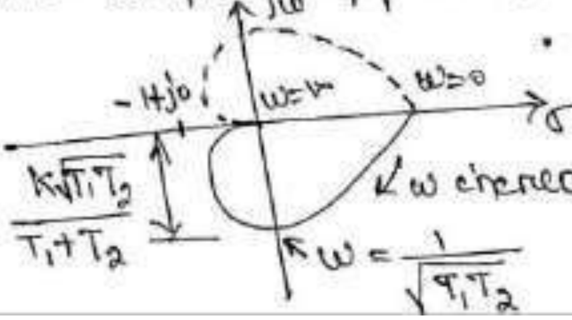
$$\phi = \angle G(j\omega)H(j\omega) = -\tan^{-1}(\omega T_1) - \tan^{-1}(\omega T_2)$$

ω	M	φ
0	K	0°
∞	0	−180°



Step 3: Path co is the mirror image of path oa shown by a doted line in the figure as given below.

Step 4: The complete nyquist plot is:

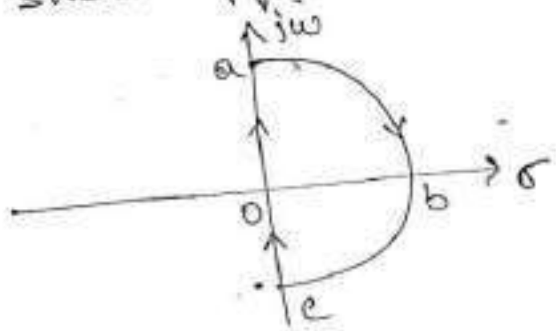


Step 5: It is seen that the plot of $G(j\omega)$ does not encircle the point $(-1+j0)$ i.e $N=0$, therefore the system is stable.

Example 2: Sketch Nyquist plot for the given OLTF.

$$G(s)H(s) = \frac{s+2}{(s+1)(s-1)}$$

Soln Step 1: Sketch Nyquist path for the OLTF.



Step 2: For path oa put $s=j\omega$ & sketch polar plot.

$$G(j\omega)H(j\omega) = \frac{2+j\omega}{(1+j\omega)(-1+j\omega)}$$

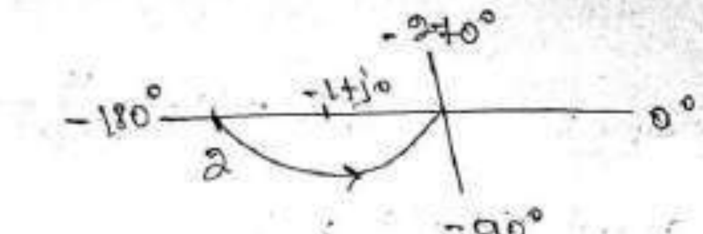
$$M = |G(j\omega)H(j\omega)| = \frac{\sqrt{4+\omega^2}}{\sqrt{1+\omega^2}\sqrt{1+\omega^2}}$$

$$\phi = \angle G(j\omega)H(j\omega) = \tan^{-1}\frac{\omega}{2} - \tan^{-1}\frac{\omega}{1} - \tan^{-1}(-\omega)$$

$$= \tan^{-1}\frac{\omega}{2} - \tan^{-1}\omega - (180 - \tan^{-1}\omega)$$

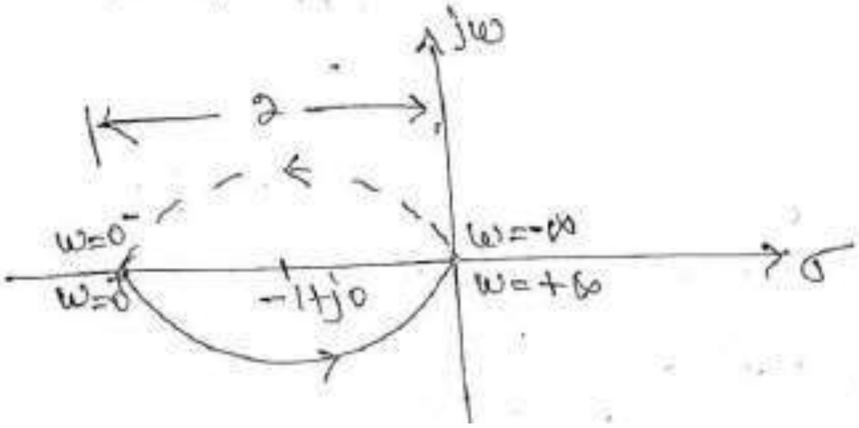
ω	M	ϕ
0	2	-180°
∞	0	-90°

Polar plot is



Step 3: Path CO is the mirror image of the path OA, as shown below:

Step 4: The complete contour for $F_1(j\omega)$ is as given below.



Step 5: Check for stability

From the OLTF no. of pole present in the right half of s-plane $P=1$.

From Nyquist plot no of counter clockwise encirclement $N=1$.

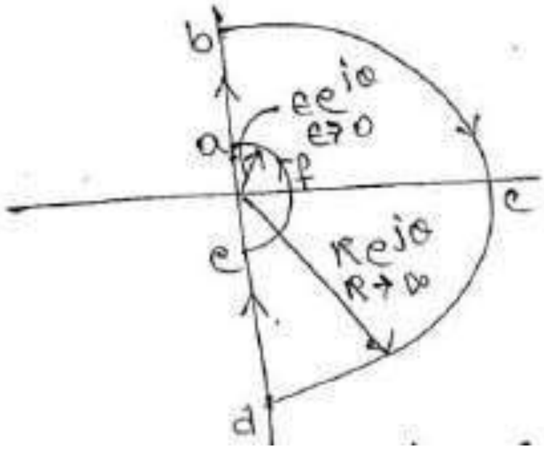
$\therefore \quad N = P - Z \Rightarrow 1 = 1 - Z \Rightarrow Z = 0$

which states that there are no zeros of $1+G(s)H(s)$ in the right half s-plane and hence the closed-loop system is stable.

Example 3 $\qquad G(s)H(s) = \dfrac{K}{s(Ts+1)}$

Sketch Nyquist plot ?

Solⁿ Step 1: As the given OLTF has pole at origin the Nyquist contour is as given below:



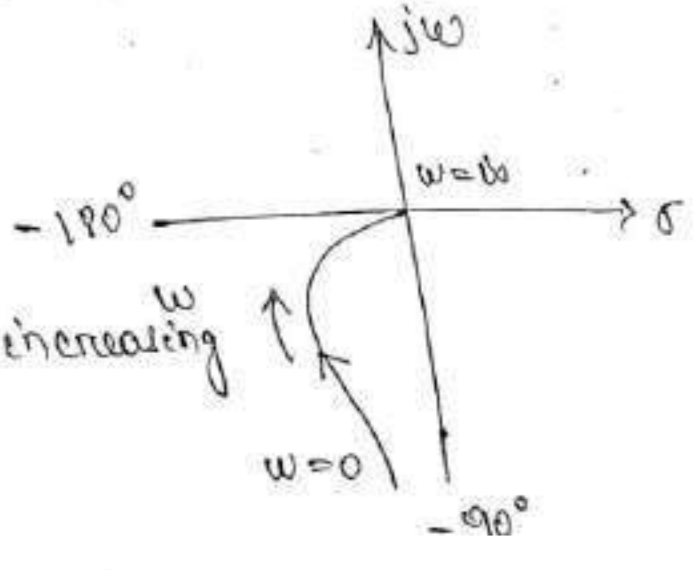
Step2: For path ab put $s = j\omega$ (where ω varies from 0 to $+\infty$) and sketch polar plot.

$$G(j\omega)H(j\omega) = \frac{K}{j\omega(j\omega T+1)}$$

$$M = |G(j\omega)H(j\omega)| = \frac{K}{\omega\sqrt{(\omega T)^2+1}}$$

$$\phi = \angle G(j\omega)H(j\omega) = -90^\circ - \tan^{-1}(\omega T)$$

$\omega = 0$	$M = \infty$	$\phi = -90^\circ$
$\omega = \infty$	$M = 0$	$\phi = -180^\circ$



Step3: For path de, it is the mirror image of path ab where ω varies from $-\infty$ to 0.

Step 4: For path 'bcd' put $s = Re^{j\theta}$ where $R \to \infty$ and θ varies from $+90°$ through $0°$ to $-90°$.

$$G(j\omega)H(j\omega)\Big|_{\substack{s=Re^{j\theta}\\ R\to\infty}} = \lim_{R\to\infty} \frac{K}{Re^{j\theta}(TRe^{j\theta}+1)}$$

$$= \lim_{R\to\infty} \frac{K}{TR^2 e^{j2\theta}} = 0e^{-j2\theta}$$

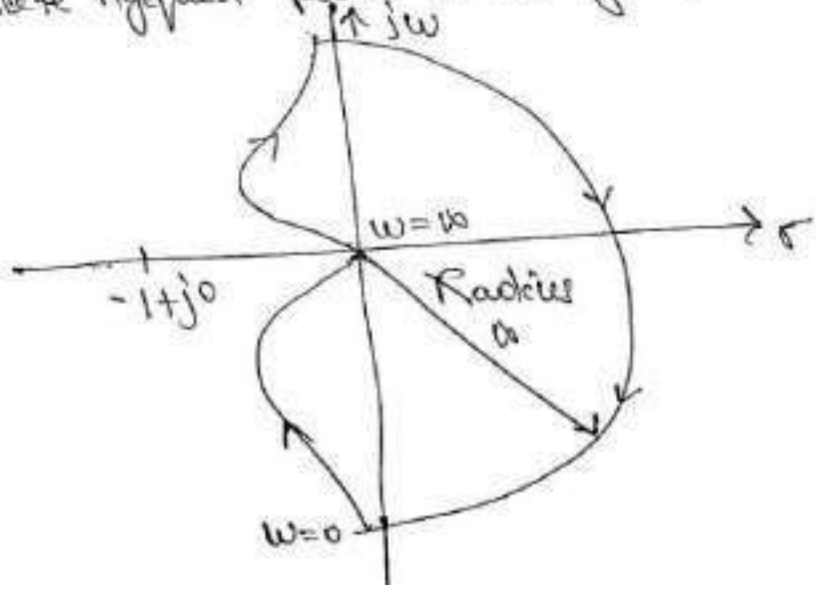
$\therefore$ $G(s)H(s)$ locus turns at the origin with zero radius from $-180°$ through $0°$ to $+180°$.

Step 5: For path 'efa' put $s = \epsilon e^{j\theta}$ where $\epsilon \to 0$ and θ varies from $-90°$ through $0°$ to $+90°$.

$$G(j\omega)H(j\omega)\Big|_{\substack{s=\epsilon e^{j\theta}\\ \epsilon\to 0}} = \lim_{\epsilon\to 0} \frac{K}{\epsilon e^{j\theta}(T\epsilon e^{j\theta}+1)}$$

$$= \lim_{\epsilon\to 0} \frac{K}{\epsilon e^{j\theta}} = \infty e^{-j\theta}$$

Thus $G(s)H(s)$ locus turns at the origin with ∞ radius from $+90°$ through $0°$ to $-90°$. Hence the complete nyquist path is as given below:



Step-6: Check for stability:

As P= 0; N=0; Hence Z=0, System is stable.

DIGITAL ELECTRONICS AND MICROPROCESSOR
5TH SEMESTER ELECTRICAL ENGINEERING

UNIT-1BASICS OF DIGITAL ELECTRONICS

Introduction to Digital Electronics:

- Digital electronics deals with the electronic manipulation of numbers, or with the manipulation of varying quantities by means of numbers.
- Because it is convenient to do so, today's digital systems deal only with the numbers 'zero' and 'one', because they can be represented easily by 'off and 'on' within a circuit.
- This is not the limitation it might seem, for the binary system of counting can be used to represent any number that we can represent with the usual decimal (0 to 9) system that we use in everyday life.
- Digital Electronics is very important in today's life because if digital circuits compared to analog circuits are that signals represented digitally can be transmitted without degradation due to noise.

Advantages of Digital Circuits

- High accuracy and programmability
- Storage of digital data is easy
- Immune to noise
- Can be implemented in the form of integrated circuits (ICs)
- Greater reliability and flexibility

Disadvantages of Digital Circuits

- Expensive
- Operate on digital signals only
- Complex circuitry

Applications of Digital Circuits

- Mobile Phones, Calculators and Digital Computers
- Radios and communication Devices
- Signal Generator
- Smart Card
- Cathode Ray Oscilloscope (CRO)
- Analog to digital converters (ADC)
- Digital to analog converters (DAC), etc.

Number system:

- A number system is defined as a system of writing to express numbers.
- It is the mathematical notation for representing numbers of a given set by using digits or other symbols in a consistent manner.
- It provides a unique representation of every number and represents the arithmetic and algebraic structure of the figures.
- It also allows us to operate arithmetic operations like addition, subtraction and division. The value of any digit in a number can be determined by:
- The digit
- Its position in the number
- The base of the number system **Types of number system:**

There are various types of number systems in mathematics. The four most common number system types are:

- Decimal number system (Base- 10)
- Binary number system (Base- 2)
- Octal number system (Base-8)

- Hexadecimal number system (Base- 16) A number N in base or radix ‘r’ can be written as:

$(N)_b = d_{n-1}\ d_{n-2}$ -- -- -- -- $d_1\ d_0\ .\ d_{-1}\ d_{-2}$ ---------------------- d_{-m}
In the above, dn-1 to d0 is the integer part, then follows a radix point, and then d-1 to d-m is the fractional part.
D_{n-1} = Most significant bit (MSB) d_{-m}
= Least significant bit (LSB)

Decimal number system:

The base or radix of Decimal number system is 10. So, the numbers ranging from 0 to 9 are used in this number system. Mathematically, we can write it as
$1358.246 = (1 \times 10^3) + (3 \times 10^2) + (5 \times 10^1) + (8 \times 10^0) + (2 \times 10^{-1}) +(4 \times 10^{-2}) + (6 \times 10^{-3})$

Binary number system:

All digital circuits and systems use this binary number system. The base or radix of this number system is 2. So, the numbers 0 and 1 are used in this number system.
Mathematically, we can write it as
$1101.011 = (1 \times 2^3) + (1 \times 2^2) + (0 \times 2^1) + (1 \times 2^0) + (0 \times 2^{-1}) +(1 \times 2^{-2}) + (1 \times 2^{-3})$

Octal number system:

The base or radix of octal number system is 8. So, the numbers ranging from 0 to 7 are used in this number system.
Mathematically, we can write it as
$1457.236 = (1 \times 8^3) + (4 \times 8^2) + (5 \times 8^1) + (7 \times 8^0) + (2 \times 8^{-1}) +(3 \times 8^{-2}) + (6 \times 8^{-3})$

Hexadecimal number system:

The base or radix of Hexa-decimal number system is 16. So, the numbers ranging from 0 to 9 and the letters from A to F are used in this number system. The decimal equivalent of Hexadecimal digits from A to F are 10 to 15.
Mathematically, we can write it as
$1A05.2C4 = (1 \times 16^3) + (10 \times 16^2) + (0 \times 16^1) + (5 \times 16^0) + (2 \times 16^{-1}) +(12 \times 16^{-2}) + (4 \times 16^{-3})$

Conversion from one system to another number system:

Decimal number system to other number system:

If the decimal number contains both integer part and fractional part, then convert both the parts of decimal number into another base individually. Steps for converting the decimal number into its equivalent number of any base ‘r’-

- Do division of integer part of decimal number and **successive quotients** with base ‘r’ and note down the remainders till the quotient is zero. Consider the remainders in reverse order to get the integer part of equivalent number of base ‘r’. That means, first and last remainders denote the least significant digit and most significant digit respectively.
- Do multiplication of fractional part of decimal number and **successive fractions** with base ‘r’ and note down the carry till the result is zero or the desired number of equivalent digits is obtained. Consider the normal sequence of carry in order to get the fractional part of equivalent number of base ‘r’.

Decimal to binary:

Example- $(152.25)_{10}$Step 1:

Divide the number 152 and its successive quotients with base 2.

Operation	Quotient	Remainder

152/2	76	0 (LSB)
76/2	38	0
38/2	19	0
19/2	9	1
9/2	4	1
4/2	2	0
2/2	1	0
1/2	0	1(MSB)

$(152)_{10} = (10011000)_2$ Step 2:

Now, perform the multiplication of 0.27 and successive fraction with base 2.

Operation	Result	carry
0.25×2	0.50	0
0.50×2	0	1

$(0.25)_{10} = (.01)_2$

Decimal to octal:

Example- $(152.25)_{10}$Step 1:

Divide the number 152 and its successive quotients with base 8.

Operation	Quotient	Remainder
152/8	19	0
19/8	2	3
2/8	0	2

$(152)_{10} = (230)_8$

Step 2:

Now perform the multiplication of 0.25 and successive fraction with base 8.

Operation	Result	carry
0.25×8	**0**	**2**

$(0.25)_{10}=(2)_8$

So, the octal number of the decimal number 152.25 is **230.2**

Decimal to hexadecimal:

Example- $(152.25)_{10}$ Step 1:

Divide the number 152 and its successive quotients with base 8.

Operation	Quotient	Remainder
152/16	9	8
9/16	0	9

$(152)_{10} = (98)_{16}$

Step 2:

Now perform the multiplication of 0.25 and successive fraction with base 16.

Operation	Result	carry
0.25×16	0	4

$(0.25)_{10} = (4)_{16}$

So, the hexadecimal number of the decimal number 152.25 is **230.4.**

Binary to other number system:

Binary to decimal:

The process starts from multiplying the bits of binary number with its corresponding positional weights. And lastly, we add all those products.

Example- $(10110.001)_2$

$(10110.001)_2$ $=(1\times2^4)+(0\times2^3)+(1\times2^2)+(1\times2^1)+(0\times2^0)+(0\times2^{-1})+(0\times2^{-2})+(1\times2^{-3})$

$(10110.001)_2$ $=(1\times16)+(0\times8)+(1\times4)+(1\times2)+(0\times1)+(0\times1/2)+(0\times1/4)+(1\times1/8)$

$(10110.001)_2$ $=16+0+4+2+0+0+0+0.125$

$(10110.001)_2=(22.125)_{10}$

Binary to octal:

In a binary number, the pair of three bits is equal to one octal digit. Two steps to convert a binary number into an octal number which are as follows:

- In the first step, we have to make the pairs of three bits on both sides of the binary point. If there will be one or two bits left in a pair of three bits pair, we add the required number of zeros on extreme sides.
- In the second step, we write the octal digits corresponding to each pair. **Example- $(111110101011.0011)_2$**

1. Firstly, we make pairs of three bits on both sides of the binary point. 111 110 101 011.001 1

On the right side of the binary point, the last pair has only one bit. To make it a complete pair of three bits, we added two zeros on the extreme side.

111 110 101 011.001 100 2. Then, we wrote the octal digits, which correspond to each pair.

$(111110101011.0011)_2$ = $(7653.14)_8$ **Binary to hexadecimal:**

The base numbers of binary and hexadecimal are 2 and 16, respectively. In a binary number, the pair of four bits is equal to one hexadecimal digit. There are also only two steps to convert a binary number into a hexadecimal number which are as follows:

1. In the first step, we have to make the pairs of four bits on both sides of the binary point. If there will be one, two, or three bits left in a pair of four bits pair, we add the required number of zeros on extreme sides.
2. In the second step, we write the hexadecimal digits corresponding to each pair. **Example- $(10110101011.0011)_2$**

1. Firstly, we make pairs of four bits on both sides of the binary point.

111 1010 1011.0011

On the left side of the binary point, the first pair has three bits. To make it a complete pair of four bits, add one zero on the extreme side.

0111 1010 1011.0011

2. Then, we write the hexadecimal digits, which correspond to each pair.

$(011110101011.0011)_2$=$(7AB.3)_{16}$Octal to other number system:

Octal to decimal:

The process starts from multiplying the digits of octal numbers with its corresponding positional weights. And lastly, we add all those products.

Example- $(152.25)_8$ Step 1:

We multiply each digit of **152.25** with its respective positional weight, and last, we add the products of all the bits with its weight.

$(152.25)_8$ $=(1\times8^2)+(5\times8^1)+(2\times8^0)+(2\times8^{-1})+(5\times8^{-2})$

$(152.25)_8$=64+40+2+(2×1/8)+(5×1/64)

$(152.25)_8$=64+40+2+0.25+0.078125

$(152.25)_8$=106.328125

So, the decimal number of the octal number 152.25 is **106.328125Octal to binary:**

The process of converting octal to binary is the reverse process of binary to octal. We write the three bits binary code of each octal number digit. **Example- $(152.25)_8$**

We write the three-bit binary digit for 1, 5, 2, and 5.

$(152.25)_8$ = $(001101010.010101)_2$

So, the binary number of the octal number 152.25 is **$(001101010.010101)_2$Octal to hexadecimal:**

For converting octal to hexadecimal, there are two steps required to perform, which are as follows:

1. In the first step, we will find the binary equivalent of number.
2. Next, we have to make the pairs of four bits on both sides of the binary point. If there will be one, two, or three bits left in a pair of four bits pair, we add the required number of

zeros on extreme sides and write the hexadecimal digits corresponding to each pair. **Example- $(152.25)_8$Step 1:**

We write the three-bit binary digit for 1, 5, 2, and 5.

$(152.25)_8$ =$(001101010.010101)_2$

So, the binary number of the octal number 152.25 is **$(001101010.010101)_2$Step 2:**

1. Now, we make pairs of four bits on both sides of the binary point.

0 0110 1010.0101 01

On the left side of the binary point, the first pair has only one digit, and on the right side, the last pair has only two-digit. To make them complete pairs of four bits, add zeros on extreme sides.

0000 0110 1010.0101 0100

2. Now, we write the hexadecimal digits, which correspond to each pair.

$(0000\quad 0110\quad 1010.0101\quad 0100)_2$ = $(6A.54)_{16}$

Hexadecimal to other number system:

Hexadecimal to decimal:

The process of converting hexadecimal to decimal is the same as binary to decimal. The process starts from multiplying the digits of hexadecimal numbers with its corresponding positional weights. And lastly, we add all those products.

Let's take an example to understand how the conversion is done from hexadecimal to decimal.

Example- $(152A.25)_{16}$Step 1:

We multiply each digit of **152A.25** with its respective positional weight, and last we add the products of all the bits with its weight.

$(152A.25)_{16}=(1\times16^3)+(5\times16^2)+(2\times16^1)+(A\times16^0)+(2\times16^{-1})+(5\times16^{-2})$

$(152A.25)_{16}=(1\times4096)+(5\times256)+(2\times16)+(10\times1)+(2\times16^{-1})+(5\times16^{-2})$

$(152A.25)_{16}=4096+1280+32+10+(2\times1/16)+(5\times1/256)$

$(152A.25)_{16}=5418+0.125+0.125$

$(152A.25)_{16}=5418.14453125$

So, the decimal number of the hexadecimal number 152A.25 is **5418.14453125**

Hexadecimal to binary:

The process of converting hexadecimal to binary is the reverse process of binary to hexadecimal. We write the four bits binary code of each hexadecimal number digit. **Example - $(152A.25)_{16}$**

We write the four-bit binary digit for 1, 5, A, 2, and 5.

$(152A.25)_{16}$ = $(0001\ 0101\ 0010\ 1010.0010\ 0101)_2$

So, the binary number of the hexadecimal number 152.25 is **$(1010100101010.00100101)_2$**

Binary equivalent	Hexadecimal
0000	0
0001	1
0010	2
0011	3
0100	4
0101	5
0110	6
0111	7
1000	8
1001	9
1010	A
1011	B
1100	C
1101	D
1110	E
1111	F

Hexadecimal to octal:

For converting hexadecimal to octal, there are two steps required to perform, which are as follows:

1. In the first step, we will find the binary equivalent of the hexadecimal number.
2. Next, we have to make the pairs of three bits on both sides of the binary point. If there will be one or two bits left in a pair of three bits pair, we add the required number of zeros on extreme sides and write the octal digits corresponding to each pair.

Example- $\mathbf{(152A.25)_{16}}$**Step 1:**

We write the four-bit binary digit for 1, 5, 2, A, and 5.

$\mathbf{(152A.25)_{16} = (0001\ 0101\ 0010\ 1010.0010\ 0101)_2}$

So, the binary number of hexadecimal number 152A.25 is $\mathbf{(0011010101010.010101)_2}$**Step 2:**

3. Then, we make pairs of three bits on both sides of the binary point.

001 010 100 101 010.001 001 010

4. Then, we write the octal digit, which corresponds to each pair.

$\mathbf{(001010100101010.001001010)_2 = (12452.112)_8}$

So, the octal number of the hexadecimal number 152A.25 is **12452.112**

Arithmetic operation:

Two types of operation that are performed on binary data include arithmetic and logic operations. Basic arithmetic operations include addition, subtraction, multiplication and division.

Binary addition:

There are four rules for binary addition:

Input A	Input B	Sum (S) A+B	Carry (C)
0	0	0	0
0	1	1	0
1	0	1	0
1	1	0	1

Example-

0011010 + 001100 = 00100110

```
  1 1        carry
  0011010   = 26₁₀
 +0001100   = 12₁₀
 --------
  0100110   = 38₁₀
```

Binary subtraction:

There are four rules for binary subtraction:

Input A	Input B	Subtract (S) A-B	Borrow (B)
0	0	0	0
0	1	0	1
1	0	1	0
1	1	0	0

Example-

0011010 - 001100 = 00001110

```
    1 1      borrow
  0011010   = 26₁₀
 -0001100   = 12₁₀
 --------
  0001110   = 14₁₀
```

Binary multiplication:

There are four rules of binary multiplication.

Input A	Input B	Multiply (M) AxB
0	0	0
0	1	0
1	0	0
1	1	1

Example:

0011010 x 001100 = 100111000

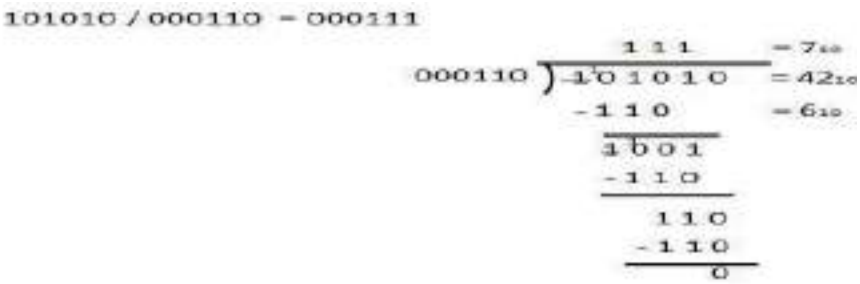
```
       0011010   = 26₁₀
      x0001100   = 12₁₀
    ----------
       0000000
      0000000
     0011010
    0011010
    ----------
    0100111000   = 312₁₀
```

Binary division:

There are four parts in any division: Dividend, Divisor, quotient, and remainder.

Input A	Input B	Divide (D) A/B
0	0	Not defined
0	1	0
1	0	Not defined
1	1	1

Example-



Signed binary number representation:

- In mathematics, positive numbers (including zero) are represented as unsigned numbers.
- That is, we do not put the +ve sign in front of them to show that they are positive numbers.
- However, when dealing with negative numbers we do use a -ve sign in front of the number to show that the number is negative in value and different from a positive unsigned value, and the same is true with signed binary numbers.
- However, in digital circuits there is no provision made to put a plus or even a minus sign to a number, since digital systems operate with binary numbers that are represented in terms of “0’s” and “1’s”.
- For signed binary numbers the most significant bit (MSB) is used as the sign bit.
- If the sign bit is “0”, this means the number is positive in value.
- If the sign bit is “1”, then the number is negative in value.

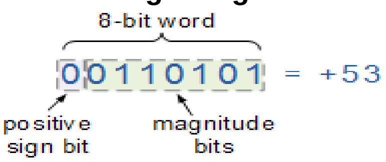
3 ways to represent negative binary number-

1. Sign magnitude
2. 1’s compliment
3. 2’ compliment **Sign magnitude:**

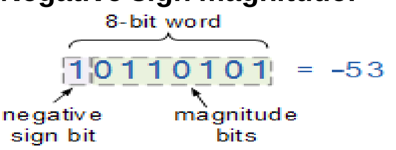
Left most digit is used to indicate the sign and the remaining digits the magnitude or value of the number.

Example-

Positive sign magnitude:



Negative sign magnitude:



The disadvantage here is that whereas before we had a full range n-bit unsigned binary number, we now have an n-1 bit signed binary number giving a reduced range of digits from:

$$-2_{(n-1)} \quad \text{to} \quad +2_{(n-1)}$$

1’s complement:

- The one’s complement of a negative binary number is the complement of its positive counterpart.
- Thus, the one’s complement of “1” is “0” and vice versa, then the one’s complement of 100101002 is simply 011010112 as all the 1’s are changed to 0’s and the 0’s to 1’s.

For representing the positive numbers, there is nothing to do.

- But for representing negative numbers, we have to use 1's complement technique.
- For representing the negative number, we first have to represent it with a positive sign, and then we find the 1's complement of it.

Example- 11010.1101

For finding 1's complement of the given number, change all 0's to 1 and all 1's to 0. So, the 1's complement of the number 11010.1101 comes out 00101.0010. **2's complement:**

- 2's complement is also used to represent the signed binary numbers.
- For finding 2's complement of the binary number, we will first find the 1's complement of the binary number and then add 1 to the least significant bit of it.

Example- 110100

For finding 2's complement of the given number, change all 0's to 1 and all 1's to 0. So, the 1's complement of the number 110100 is 001011. Now add 1 to the LSB of this number, i.e., (001011)+1=001100.

Addition and subtraction using 1's complement:

There are three different cases possible when we add two binary numbers which are as follows:

Case 1: Addition of the positive number with a negative number when the positive number has a greater magnitude.

Initially, calculate the 1's complement of the given negative number. Sum up with the given positive number. If we get the end-around carry 1, it gets added to the LSB.

Example: 1101 and -1001

- First, find the 1's complement of the negative number 1001. So, for finding 1's complement, change all 0 to 1 and all 1 to 0. The 1's complement of the number 1001 is 0110.
- Now, add both the numbers, i.e., 1101 and 0110; 1101+0110=1 0011
- By adding both numbers, we get the end-around carry 1. We add this end around carry to the LSB of 0011.

  0011+1=0100

Case 2: Adding a positive value with a negative value in case the negative number has a higher magnitude.

Initially, calculate the 1's complement of the negative value. Sum it with a positive number. In this case, we did not get the end-around carry. So, take the 1's complement of the result to get the final result.

Note: The resultant is a negative value.

Example: 1101 and -1110

- First find the 1's complement of the negative number 1110. So, for finding 1's complement, we change all 0 to 1, and all 1 to 0. 1's complement of the number 1110 is 0001.
- Now, add both the numbers, i.e., 1101 and 0001; 1101+0001= 1110
- Now, find the 1's complement of the result 1110 that is the final result. So, the 1's complement of the result 1110 is 0001, and we add a negative sign before the number so that we can identify that it is a negative number.

Case 3: Addition of two negative numbers

In this case, first find the 1's complement of both the negative numbers, and then we add both these complement numbers. In this case, we always get the end-around carry, which get added to the LSB, and for getting the final result, we take the 1's complement of the result.

Note: The resultant is a negative value.

Example: -1101 and -1110 in five-bit register

- Firstly, find the 1's complement of the negative numbers 01101 and 01110. So, for finding 1's complement, we change all 0 to 1, and all 1 to 0. 1's complement of the number 01110 is 10001, and 01101 is 10010.

- Now, we add both the complement numbers, i.e., 10001 and 10010; 10001+10010= 1 00011
- By adding both numbers, we get the end-around carry 1. We add this end-around carry to the LSB of 00011.
  00011+1=00100
- Now, find the 1's complement of the result 00100 that is the final answer. So, the 1's complement of the result 00100 is 110111, and add a negative sign before the number so that we can identify that it is a negative number.

Addition and subtraction using 2's complement:

There are three different cases possible when we add two binary numbers using 2's complement, which is as follows:

Case 1: Addition of the positive number with a negative number when the positive number has a greater magnitude.

Initially find the 2's complement of the given negative number. Sum up with the given positive number. If we get the end-around carry 1 then the number will be a positive number and the carry bit will be discarded and remaining bits are the final result.

Example: 1101 and -1001

- First, find the 2's complement of the negative number 1001. So, for finding 2's complement, change all 0 to 1 and all 1 to 0 or find the 1's complement of the number 1001. The 1's complement of the number 1001 is 0110, and add 1 to the LSB of the result 0110. So the 2's complement of number 1001 is 0110+1=0111
- Add both the numbers, i.e., 1101 and 0111;
  1101+0111=1 0100
- By adding both numbers, we get the end-around carry 1. We discard the end-around carry. So, the addition of both numbers is 0100.

Case 2: Adding of the positive value with a negative value when the negative number has a higher magnitude.

Initially, add a positive value with the 2's complement value of the negative number. Here, no end-around carry is found. So, we take the 2's complement of the result to get the final result.

Note: The resultant is a negative value.

Example: 1101 and -1110

- First, find the 2's complement of the negative number 1110. So, for finding 2's complement, add 1 to the LSB of its 1's complement value 0001.
  0001+1=0010
- Add both the numbers, i.e., 1101 and 0010; 1101+0010= 1111
- Find the 2's complement of the result 1110 that is the final result. So, the 2's complement of the result 1110 is 0001, and add a negative sign before the number so that we can identify that it is a negative number.

Case 3: Addition of two negative numbers

In this case, first, find the 2's complement of both the negative numbers, and then we will add both these complement numbers. In this case, we will always get the end-around carry, which will be added to the LSB, and forgetting the final result, we will take the2's complement of the result.

Note: The resultant is a negative value.

Example: -1101 and -1110 in five-bit register

- Firstly, find the 2's complement of the negative numbers 01101 and 01110. So, for finding 2's complement, we add 1 to the LSB of the 1's complement of these numbers. 2's complement of the number 01110 is 10010, and 01101 is 10011.
- We add both the complement numbers, i.e., 10001 and 10010; 10010+10011= 1 00101

- By adding both numbers, we get the end-around carry 1. This carry is discarded and the final result is the 2.s complement of the result 00101. So, the 2's complement of the result 00101 is 11011, and we add a negative sign before the number so that we can identify that it is a negative number.

Digital codes:

In the coding, when numbers or letters are represented by a specific group of symbols, it is said to be that number or letter is being encoded. The group of symbols is called as code. The digital data is represented, stored and transmitted as group of bits. This group of bits is also called as binary code.

Advantages of Binary Code:

- Binary codes are suitable for the computer applications.
- Binary codes are suitable for the digital communications.
- Binary codes make the analysis and designing of digital circuits if we use the binary codes.
- Since only 0 & 1 are being used, implementation becomes easy.

Classification of binary codes:

The codes are broadly categorized into following four categories.

- Weighted Codes
- Non-Weighted Codes
- Binary Coded Decimal Code
- Alphanumeric Codes
- Error Detecting Codes
- Error Correcting Codes **Weighted Codes:**

Weighted binary codes are those binary codes which obey the positional weight principle. Each position of the number represents a specific weight. Several systems of the codes are used to express the decimal digits 0 through 9. In these codes each decimal digit is represented by a group of four bits.

Decimal Digit	8421 Code	2421 Code	84-2-1 Code
0	0000	0000	0000
1	0001	0001	0111
2	0010	0010	0110
3	0011	0011	0101
4	0100	0100	0100
5	0101	1011	1011

6	0110	1100	1010
7	0111	1101	1001
8	1000	1110	1000
9	1001	1111	1111

8 4 2 1 code

- The weights of this code are 8, 4, 2 and 1.
- This code has all positive weights. So, it is a **positively weighted code**.
- This code is also called as **natural BCDBinary Coded Decimalcode**.
- In this code each decimal digit is represented by a 4-bit binary number.
- BCD is a way to express each of the decimal digits with a binary code.
- In the BCD, with four bits we can represent sixteen numbers (0000 to 1111).
- But in BCD code only first ten of these are used (0000 to 1001).
- The remaining six code combinations i.e., 1010 to 1111 are invalid in BCD.

Decimal	0	1	2	3	4	5	6	7	8	9
BCD	0000	0001	0010	0011	0100	0101	0110	0111	1000	1001

Example

Let us find the BCD equivalent of the decimal number 786. This number has 3 decimal digits 7, 8 and 6. From the table, we can write the BCD 84218421 codes of 7, 8 and 6 are 0111, 1000 and 0110 respectively.

$$786_{10} = 0111\ 1000\ 0110_{BCD}$$

There are 12 bits in BCD representation, since each BCD code of decimal digit has 4 bits.

Advantages of BCD Codes

- It is very similar to decimal system.
- We need to remember binary equivalent of decimal numbers 0 to 9 only.

Disadvantages of BCD Codes

- The addition and subtraction of BCD have different rules.
- The BCD arithmetic is little more complicated.
- BCD needs more number of bits than binary to represent the decimal number. So, BCD is less efficient than binary.

Non-weighted code:

n this type of binary codes, the positional weights are not assigned. The examples of nonweighted codes are Excess-3 code and Gray code.

Excess-3 code

- The Excess-3 code is also called as XS-3 code.
- It is non-weighted code used to express decimal numbers.

- The Excess-3 code words are derived from the 8421 BCD code words adding (0011)2 or (3)10 to each code word in 8421.
- The excess-3 codes are obtained as follows –

$$\text{Decimal Number} \longrightarrow \text{8421 BCD} \xrightarrow[\text{0011}]{\text{Add}} \text{Excess-3}$$

Decimal	BCD 8 4 2 1	Excess-3 BCD + 0011
0	0 0 0 0	0 0 1 1
1	0 0 0 1	0 1 0 0
2	0 0 1 0	0 1 0 1
3	0 0 1 1	0 1 1 0
4	0 1 0 0	0 1 1 1
5	0 1 0 1	1 0 0 0
6	0 1 1 0	1 0 0 1
7	0 1 1 1	1 0 1 0
8	1 0 0 0	1 0 1 1
9	1 0 0 1	1 1 0 0

Gray Code

- It is the non-weighted code and it is not arithmetic codes.
- That means there are no specific weights assigned to the bit position.
- It has a very special feature that, only one bit will change each time the decimal number is incremented as shown in fig.
- As only one-bit changes at a time, the gray code is called as a unit distance code.
- The gray code is a cyclic code. Gray code cannot be used for arithmetic operation.

Decimal	BCD	Gray
0	0 0 0 0	0 0 0 0
1	0 0 0 1	0 0 0 1
2	0 0 1 0	0 0 1 1
3	0 0 1 1	0 0 1 0
4	0 1 0 0	0 1 1 0
5	0 1 0 1	0 1 1 1
6	0 1 1 0	0 1 0 1
7	0 1 1 1	0 1 0 0
8	1 0 0 0	1 1 0 0
9	1 0 0 1	1 1 0 1

Binary code to Gray Code Conversion:

- Consider the given binary code and place the MSB of binary to the left of MSB.

- Compare the successive two bits starting from MSB. If the 2 bits are same, then the output is zero. Otherwise, output is one.
- Repeat the above step till the LSB of Gray code is obtained.

Example-

From the table, we know that the Gray code corresponding to binary code 1000 is 1100. Now, let us verify it by using the above procedure. Given, binary code is 1000. Step 1 − By placing same MSB to the left of MSB, the binary code will be 1000.

Step 2 − By comparing successive two bits of new binary code, we will get the gray code as 1100.

Application of Gray code:

- Gray code is popularly used in the shaft position encoders.
- A shaft position encoder produces a code word which represents the angular position of the shaft.

Alphanumeric codes:

- A binary digit or bit can represent only two symbols as it has only two states '0' or '1'.
- But this is not enough for communication between two computers because there we need many more symbols for communication.
- These symbols are required to represent 26 alphabets with capital and small letters, numbers from 0 to 9, punctuation marks and other symbols.
- The alphanumeric codes are the codes that represent numbers and alphabetic characters.
- Mostly such codes also represent other characters such as symbol and various instructions necessary for conveying information.
- An alphanumeric code should at least represent 10 digits and 26 letters of alphabet i.e. total 36 items.
- The following two alphanumeric codes are very commonly used for the data representation.

    i. American Standard Code for Information Interchange (ASCII).
    ii. Extended Binary Coded Decimal Interchange Code (EBCDIC).

- ASCII code is a 7-bit code whereas EBCDIC is an 8-bit code.
- ASCII code is more commonly used worldwide while EBCDIC is used primarily in large IBM computers.

Error detection codes:

- Error detection codes are used to detect the errors present in the received data bitstream.
- These codes contain some bits, which are included appended to the original bit stream.
- These codes detect the error, if it is occurred during transmission of the original data bitstream.
- Example − Parity code, Hamming code.

Error correction codes:

- Error correction codes are used to correct the errors present in the received data bitstream so that, we will get the original data.
- Error correction codes also use the similar strategy of error detection codes.
- Example − Hamming code.

Therefore, to detect and correct the errors, additional bits are appended to the data bits at the time of transmission. **Logic gates:**

- Logic gates play an important role in circuit design and digital systems.
- It is a building block of a digital system and an electronic circuit that always have only one output.

- These gates can have one input or more than one input, but most of the gates have two inputs.

We can classify these Logic gates into the following three categories.
1. Basic gates
2. Universal gates
3. Special gates **Basic gates:**

The basic gates are AND, OR & NOT gates.

AND gate:

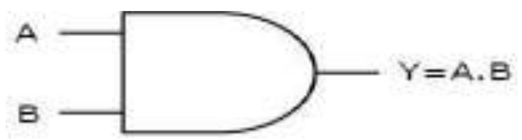
An AND gate is a digital circuit that has two or more inputs and produces an output, which is the **logical AND** of all those inputs. It is optional to represent the **Logical AND** with the symbol '.'.

The following table shows the **truth table** of 2-input AND gate.

A	B	Y = A.B
0	0	0
0	1	0
1	0	0
1	1	1

Here A, B are the inputs and Y is the output of two input AND gate. If both inputs are '1', then only the output, Y is '1'. For remaining combinations of inputs, the output, Y is '0'.

The following figure shows the **symbol** of an AND gate, which is having two inputs A, B and one output, Y.



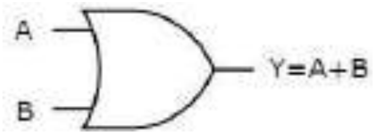
OR gate:

An OR gate is a digital circuit that has two or more inputs and produces an output, which is the logical OR of all those inputs. This **logical OR** is represented with the symbol '+'. The following table shows the **truth table** of 2-input OR gate.

A	B	Y = A + B
0	0	0
0	1	1
1	0	1
1	1	1

Here A, B are the inputs and Y is the output of two input OR gate. If both inputs are '0', then only the output, Y is '0'. For remaining combinations of inputs, the output, Y is '1'.

The following figure shows the **symbol** of an OR gate, which is having two inputs A, B and one output, Y.



This OR gate produces an output Y, which is the **logical OR** of two inputs A, B.

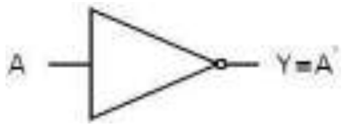
NOT gate:

A NOT gate is a digital circuit that has single input and single output. The output of NOT gate is the **logical inversion** of input. Hence, the NOT gate is also called as inverter. The following table shows the **truth table** of NOT gate.

A	Y = A’
0	1
1	0

Here A and Y are the input and output of NOT gate respectively. If the input, A is ‘0’, then the output, Y is ‘1’. Similarly, if the input, A is ‘1’, then the output, Y is ‘0’.

The following figure shows the **symbol** of NOT gate, which is having one input, A and one output, Y.



This NOT gate produces an output Y, which is the **complement** of input, A.

Universal gates

NAND & NOR gates are called as **universal gates**.

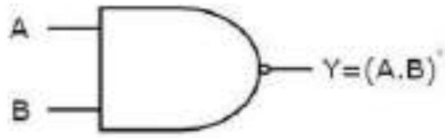
NAND gate

NAND gate is a digital circuit that has two or more inputs and produces an output, which is the **inversion of logical AND** of all those inputs.

The following table shows the **truth table** of 2-input NAND gate.

A	B	Y = (A.B)’
0	0	1
0	1	1
1	0	1
1	1	0

The following image shows the **symbol** of NAND gate, which is having two inputs A, B and one output, Y.



NAND gate operation is same as that of AND gate followed by an inverter. That's why the NAND gate symbol is represented like that.

NOR gate:

NOR gate is a digital circuit that has two or more inputs and produces an output, which is the **inversion of logical OR** of all those inputs.

The following table shows the **truth table** of 2-input NOR gate

A	**B**	**Y = (**A+B)'
0	0	1
0	1	0
1	0	0
1	1	0

The following figure shows the **symbol** of NOR gate, which is having two inputs A, B and one output, Y.

NOR gate operation is same as that of OR gate followed by an inverter. That's why the NOR gate symbol is represented like that.

Special Gates

Ex-OR & Ex-NOR gates are called as special gates. Because, these two gates are special cases of OR & NOR gates.

Ex-OR gate:

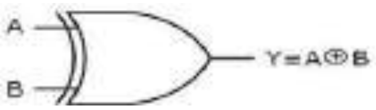
The full form of Ex-OR gate is **Exclusive-OR** gate. Its function is same as that of OR gate except for some cases, when the inputs having even number of ones. The following table shows the **truth table** of 2-input Ex-OR gate.

A	**B**	**Y = A⊕B**

0	0	0
0	1	1
1	0	1
1	1	0

The output of Ex-OR gate is ‘1’, when only one of the two inputs is ‘1’. And it is zero, when both inputs are same.

Below figure shows the **symbol** of Ex-OR gate, which is having two inputs A, B and one output, Y.



The output of Ex-OR gate is ‘1’, when odd number of ones present at the inputs. Hence, the output of Ex-OR gate is also called as an **odd function**.

Ex-NOR gate:

The full form of Ex-NOR gate is **Exclusive-NOR** gate. Its function is same as that of NOR gate except for some cases, when the inputs having even number of ones.

The following table shows the **truth table** of 2-input Ex-NOR gate.

A	**B**	**Y = A⊙B**
0	0	1
0	1	0
1	0	0
1	1	1

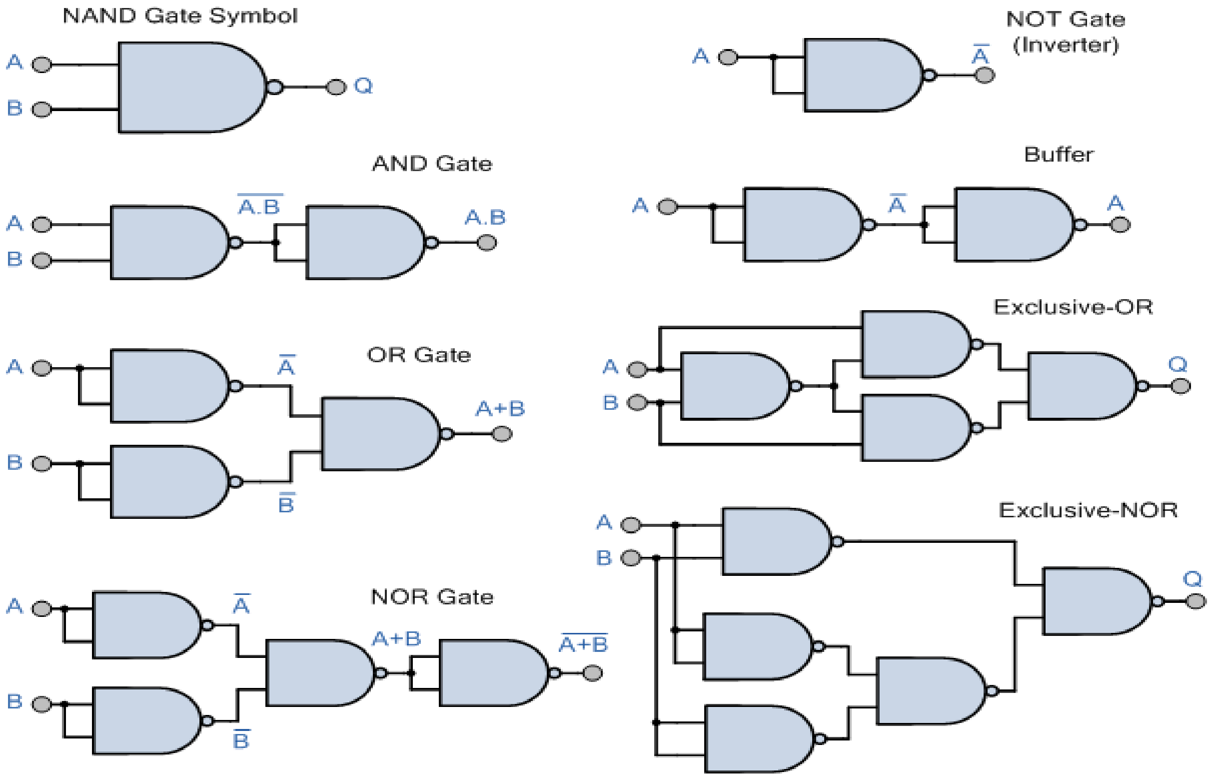
The output of Ex-NOR gate is ‘1’, when both inputs are same. And it is zero, when both the inputs are different.

The following figure shows the **symbol** of Ex-NOR gate, which is having two inputs A, B and one output, Y.

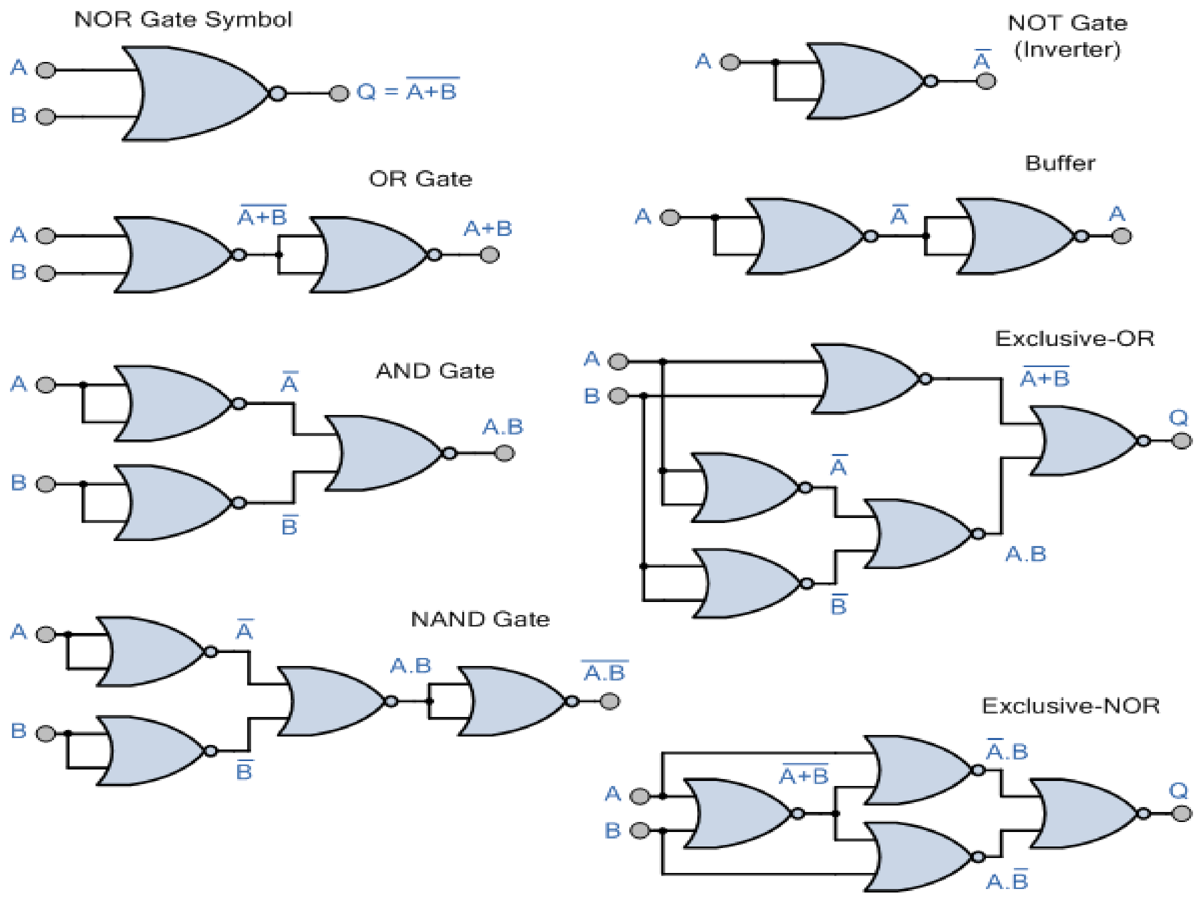
The output of Ex-NOR gate is ‘1’, when even number of ones present at the inputs. Hence, the output of Ex-NOR gate is also called as an **even function**.
From the above truth tables of Ex-OR & Ex-NOR logic gates, we can easily notice that the Ex-NOR operation is just the logical inversion of Ex-OR operation.

Universal gates and its realization:

We can realise all of the other gates by using just one single type of universal logic gate, the NAND (NOT AND) or the NOR (NOT OR) gate, thereby reducing the number of different types of logic gates required, and also the cost. Thus, the NAND and the NOR gates are commonly referred to as Universal Logic Gates. **Implementation of logic gates using NAND gate only:**



Implementation of logic gates using NOR gate only:



Boolean Algebra:

Boolean Algebra is used to analyse and simplify the digital (logic) circuits. It uses only the binary numbers i.e., 0 and 1. It is also called as Binary Algebra or logical Algebra. Boolean algebra was invented by George Boole in 1854.

Boolean Laws

There are six types of Boolean Laws.

Commutative law

Any binary operation which satisfies the following expression is referred to as commutative operation.

(i) $A.B = B.A$ (ii) $A + B = B + A$

Commutative law states that changing the sequence of the variables does not have any effect on the output of a logic circuit.

Associative law

This law states that the order in which the logic operations are performed is irrelevant as their effect is the same.

(i) $(A.B).C = A.(B.C)$ (ii) $(A + B) + C = A + (B + C)$

Distributive law

Distributive law states the following condition.

$A.(B + C) = A.B + A.C$

AND law

These laws use the AND operation. Therefore, they are called as **AND** laws.

(i) $A.0 = 0$ (ii) $A.1 = A$

(iii) $A.A = A$ (iv) $A.\overline{A} = 0$

OR law

These laws use the OR operation. Therefore, they are called as OR laws.

(i) $A + 0 = A$ (ii) $A + 1 = 1$

(iii) $A + A = A$ (iv) $A + \overline{A} = 1$

INVERSION law

This law uses the NOT operation. The inversion law states that double inversion of a variable results in the original variable itself.

$\overline{\overline{A}} = A$

Boolean Function:

Boolean algebra deals with binary variables and logic operation. A **Boolean Function** is described by an algebraic expression called **Boolean expression** which consists of binary variables, the constants 0 and 1, and the logic operation symbols. Consider the following example.

$F (A, B, C, D)$	=	$A + \overline{BC} + ADC$	Equation No. 1
Boolean Function		Boolean Expression	

Here the left side of the equation represents the output Y. So we can state equation no. 1

$Y = A + \overline{BC} + ADC$

Truth Table Formation

A truth table represents a table having all combinations of inputs and their corresponding result.

It is possible to convert the switching equation into a truth table. For example, consider the following switching equation.

$F (A, B, C) = A + BC$

The output will be high (1) if A = 1 or BC = 1 or both are 1. The truth table for this equation is shown by Table (a). The number of rows in the truth table is 2^n where n is the number of input variables (n=3 for the given equation). Hence there are $2^3 = 8$ possible input combination of inputs.

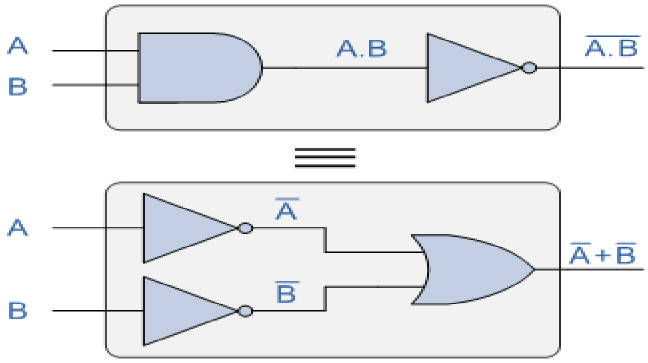
Inputs			Output
A	B	C	F
0	0	0	0
0	0	1	0
0	1	0	0
0	1	1	1
1	0	0	1
1	0	1	1
1	1	0	1
1	1	1	1

De Morgan's Theorem:

De Morgan's 1st theorem states that the complement of the product of all the terms is equal to the sum of the complement of each term.

(A.B)’ = A’ + B’

NAND

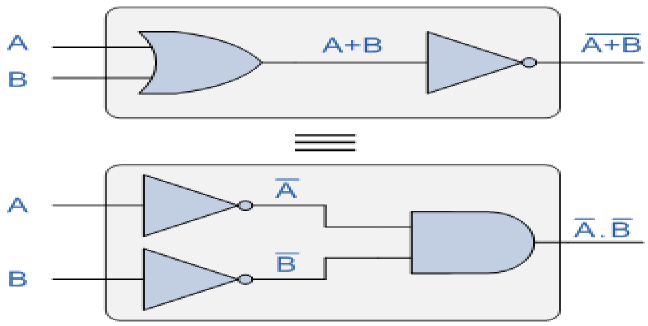


Negative-OR

De Morgan's 2nd theorem states that the complement of the sum of all the terms is equal to the product of the complement of each term.

(A + B)’ = A’. B’

NOR



Negative-AND

Duality Theorem:

This theorem states that the dual of the Boolean function is obtained by interchanging the logical AND operator with logical OR operator and zeros with ones. For every Boolean function, there will be a corresponding Dual function.

Group1	Group2
x + 0 = x	x.1 = x
x + 1 = 1	x.0 = 0
x + x = x	x.x = x
x + x’ = 1	x.x’ = 0
x + y = y + x	x.y = y.x

Example-1:
Given Boolean function, f = p’qr + pq’r + pqr’ + pqr.

Step 1 − Use the **Boolean postulate**, x + x = x. That means, the Logical OR operation with any Boolean variable ‘n’ times will be equal to the same variable. So, we can write the last term pqr two more times.

⇒ f = p’qr + pq’r + pqr’ + pqr + pqr + pqr

Step 2 − Use **Distributive law** for 1st and 4th terms, 2nd and 5th terms, 3rd and 6th terms.

⇒ f = qrp'+pp'+p + prq'+qq'+q + pqr'+rr'+r

Step 3 − Use **Boolean postulate**, x + x’ = 1 for simplifying the terms present in eachparenthesis.

⇒ f = qr11 + pr11 + pq11

Step 4 − Use **Boolean postulate**, x.1 = x for simplifying the above three terms.

⇒ f = qr + pr + pq⇒ f =

pq + qr + pr Therefore, the simplified Boolean function is **f =**

pq + qr + pr.

Example-2:
Let us find the complement **of the Boolean function, f = p’q + pq’.** The complement of Boolean function is f’ = p'q+pq'p'q+pq'’.

Step 1 − Use DeMorgan’s theorem, x+yx+y’ = x’.y’. ⇒ f’
= p'qp'q’.pq'pq'’

Step 2 − Use DeMorgan’s theorem, x.yx.y’ = x’ + y’

$\Rightarrow f' = \{pp' + q'\}.\{p' + qq'\}$

Step3 - Use the Boolean postulate, $x''=x$. $\Rightarrow f' = \{p + q'\}.\{p' + q\} \Rightarrow f' = pp' + pq + p'q' + qq'$

Step 4 − Use the Boolean postulate, $xx'=0$.

$$\Rightarrow f = 0 + pq + p'q' + 0 \Rightarrow$$

$$f = pq + p'q'$$

Therefore, the **complement** of Boolean function, p'q + pq' is **pq + p'q'**.

SOP and POS form:

Sum of Product (SOP):

- The Sum of Product expression is equivalent to the logical AND function which Sums two or more Products to produce an output.
- We will get four Boolean product terms by combining two variables x and y with logical AND operation.
- These Boolean product terms are called as **min terms** or **standard product terms**.
  - If the binary variable is '0', then it is represented as complement of variable and '1' as normal form in min term. The min terms are x'y', x'y, xy' and xy. **Product of Sum (POS):**
- The Product of Sum expression is equivalent to the logical OR-AND function which gives the AND Product of two or more OR Sums to produce an output.
- We will get four Boolean sum terms by combining two variables x and y with logical OR operation.
- These Boolean sum terms are called as **Max terms** or **standard sum terms** If the binary variable is '1', then it is represented as complement of variable and '0' as normal form in Max term. The Max terms are x + y, x + y', x' + y and x' + y'.

x	y	Min terms	Max terms
0	0	$m_0=x'y'$	$M_0=x + y$
0	1	$m_1=x'y$	$M_1=x + y'$
1	0	$m_2=xy'$	$M_2=x' + y$
1	1	$m_3=xy$	$M_3=x' + y'$

Canonical SOP and POS forms:

- A truth table consists of a set of inputs and outputs.
- If there are 'n' input variables, then there will be 2^n possible combinations with zeros and ones.

- So, the value of each output variable depends on the combination of input variables.
   So, each output variable will have ‘1’ for some combination of input variables and ‘0’ for some other combination of input variables.

Therefore, we can express each output variable in following two ways.

- Canonical SOP form
- Canonical POS form **Canonical SOP form:**
- Canonical SOP form means Canonical Sum of Products form.
- In this form, each product term contains all literals.
- So, these product terms are nothing but the min terms. Hence, canonical SOP form is also called as **sum of min terms** form.
- First, identify the min terms for which, the output variable is one and then do the logical OR of those min terms in order to get the Boolean expression function corresponding to that output variable. This Boolean function will be in the form of sum of min terms.
- Follow the same procedure for other output variables also, if there is more than one output variable. **Example**

Consider the following **truth table**.

Inputs		**Ou put**	
p	**Q**	**r**	**f**
0	0	0	0
0	0	1	0
0	1	0	0
0	1	1	1
1	0	0	0
1	0	1	1
1	1	0	1
1	1	1	1

- Here, the output f is ‘1’ for four combinations of inputs.
- The corresponding min terms are p’qr, pq’r, pqr’, pqr.
- By doing logical OR of these four min terms, we will get the Boolean function of output f.

Therefore, the Boolean function of output is, f
= p’qr + pq’r + pqr’ + pqr.

This is the **canonical SOP form** of output, f. We can also represent this function in following two notations.

$$f = m3+m5+m6+m7 \quad f=\sum m(3,5,6,7)$$

In one equation, we represented the function as sum of respective min terms. In other equation, we used the symbol for summation of those min terms. **Canonical POS form:**

- Canonical POS form means Canonical Product of Sums form.

- In this form, each sum term contains all literals. So, these sum terms are nothing but the Max terms. Hence, canonical POS form is also called as **product of Max terms** form.
- First, identify the Max terms for which, the output variable is zero and then do the logical AND of those Max terms in order to get the Boolean expression function corresponding to that output variable. This Boolean function will be in the form of product of Max terms.
- Follow the same procedure for other output variables also, if there is more than one output variable.

Example

- Consider the same truth table of previous example.
- Here, the output f is '0' for four combinations of inputs.
- The corresponding Max terms are p + q + r, p + q + r', p + q' + r, p' + q + r.
- By doing logical AND of these four Max terms, we will get the Boolean function of output f.

Therefore, the Boolean function of output is, f
= (p+q+r). (p+q+r'). (p+q'+r). (p'+q+r).
This is the **canonical POS form** of output, f.
We can also represent this function in following two notations. f=M0.M1.M2.M4

$$f=\prod M(0,1,2,4)$$

In one equation, we represented the function as product of respective Max terms. In other equation, we used the symbol for multiplication of those Max terms. The Boolean function,
f = (p+q+r). (p+q+r'). (p+q'+r). (p'+q+r)
is the dual of the Boolean function, f
= p'qr + pq'r + pqr' + pqr.
Therefore, both canonical SOP and canonical POS forms are **Dual** to each other. Functionally, these two forms are same. Based on the requirement, we can use one of these two forms.

Standard SOP and POS forms

We discussed two canonical forms of representing the Boolean outputs. Similarly, there are two standard forms of representing the Boolean outputs. These are the simplified version of canonical forms.

- Standard SOP form
- Standard POS form

The main **advantage** of standard forms is that the number of inputs applied to logic gates can be minimized. Sometimes, there will be reduction in the total number of logic gates required.

Standard SOP form:

Standard SOP form means **Standard Sum of Products** form. In this form, each product term need not contain all literals. So, the product terms may or may not be the min terms.
Therefore, the Standard SOP form is the simplified form of canonical SOP form.
We will get Standard SOP form of output variable in two steps.

- Get the canonical SOP form of output variable
- Simplify the above Boolean function, which is in canonical SOP form.

Follow the same procedure for other output variables also, if there is more than one output variable. Sometimes, it may not possible to simplify the canonical SOP form. In that case, both canonical and standard SOP forms are same.

Example

Convert the following Boolean function into Standard SOP form. f

= p'qr + pq'r + pqr' + pqr

The given Boolean function is in canonical SOP form. Now, we have to simplify this Boolean function in order to get standard SOP form.

Step 1 − Use the **Boolean postulate**, x + x = x. That means, the Logical OR operation with any Boolean variable 'n' times will be equal to the same variable. So, we can write the last term pqr two more times.

$$\Rightarrow f = p'qr + pq'r + pqr' + pqr + pqr + pqr$$

Step 2 − Use **Distributive law** for 1st and 4th terms, 2nd and 5th terms, 3rd and 6th terms.f ⇒

$$= qr\,(p'+p) + pr\,(q'+q) + pq\,(r'+r)$$

Step 3 − Use **Boolean postulate**, x + x' = 1 for simplifying the terms present in each parenthesis. ⇒f= qr 1 + pr 1 + pq 1

Step 4 − Use **Boolean postulate**, x.1 = x for simplifying above three terms.

$$\Rightarrow f = qr + pr + pq \Rightarrow$$

$$f = pq + qr + pr$$

This is the simplified Boolean function. Therefore, the **standardSOP form** corresponding to given canonical SOP form is **f = pq + qr + prStandard POS form:**

Standard POS form means **Standard Product of Sums** form. In this form, each sum term need not contain all literals. So, the sum terms may or may not be the Max terms. Therefore, the Standard POS form is the simplified form of canonical POS form.

We will get Standard POS form of output variable in two steps.

- Get the canonical POS form of output variable
- Simplify the above Boolean function, which is in canonical POS form.

Follow the same procedure for other output variables also, if there is more than one output variable. Sometimes, it may not possible to simplify the canonical POS form. In that case, both canonical and standard POS forms are same.

Example

Convert the following Boolean function into Standard POS form.

$$f = (p+q+r).\,(p+q+r').\,(p+q'+r).\,(p'+q+r)$$

The given Boolean function is in canonical POS form. Now, we have to simplify this Boolean function in order to get standard POS form.

Step 1 − Use the **Boolean postulate**, x.x = x. That means, the Logical AND operation with any Boolean variable 'n' times will be equal to the same variable. So, we can write the first term p+q+r two more times.

$$\Rightarrow f = (p+q+r).\,(p+q+r).\,(p+q+r).\,(p+q+r').\,(p+q'+r).\,(p'+q+r)$$

Step 2 − Use **Distributive law,** x + y.z = (x+y). (x+z) for 1st and 4th parenthesis, 2nd and 5th parenthesis, 3rd and 6th parenthesis.

$$\Rightarrow f = (p+q+rr').\,(p+r+qq').\,(q+r+pp')$$

Step 3 − Use **Boolean postulate**, x.x'=0 for simplifying the terms present in each parenthesis. f

$$\Rightarrow = (p+q+0).\,(p+r+0).\,(q+r+0)$$

Step 4 − Use **Boolean postulate**, x + 0 = x for simplifying the terms present in each parenthesis ⇒f= (p+q). (p+r). (q+r) f ⇒(p+q). (q+r). (p+r)

This is the simplified Boolean function. Therefore, the **standard POS form** corresponding to given canonical POS form is **f = (**p+q)**. (**q+r)**. (**p+r). This is the **dual** of the Boolean function, f = pq + qr + pr.

Therefore, both Standard SOP and Standard POS forms are Dual to each other. **Karnaugh map:**

- Karnaugh introduced a method for simplification of Boolean functions in an easy way. This method is known as Karnaugh map method or K-map method.
- It is a graphical method, which consists of 2n cells for 'n' variables. The adjacent cells are differed only in single bit position.

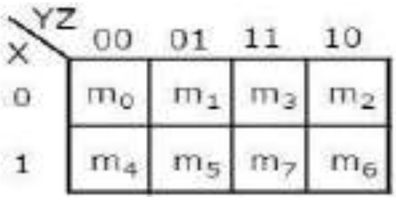
2-Variable K-Map:

The number of cells in 2 variable K-map is four, since the number of variables is two.

- There is only one possibility of grouping 4 adjacent min terms.
- The possible combinations of grouping 2 adjacent min terms are {(m_0, m_1), (m_2, m_3), (m_0, m_2) and (m_1, m_3)}. **3-Variable**

K-Map:

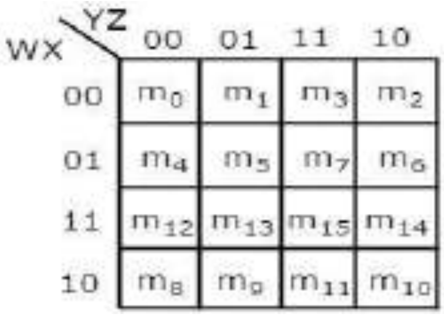
The number of cells in 3 variable K-map is eight, since the number of variables is three.

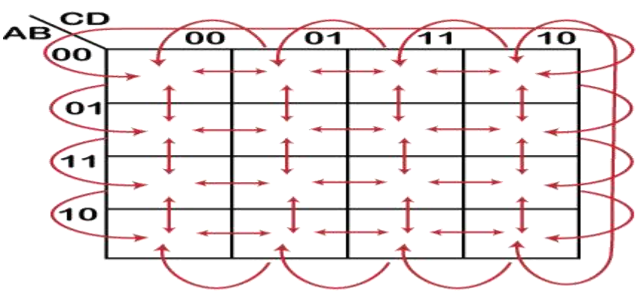


- There is only one possibility of grouping 8 adjacent min terms.
- The possible combinations of grouping 4 adjacent min terms are {(m_0, m_1, m_3, m_2), (m_4, m_5, m_7, m_6), (m_0, m_1, m_4, m_5), (m_1, m_3, m_5, m_7), (m_3, m_2, m_7, m_6) and (m_2, m_0, m_6, m_4)}.
- The possible combinations of grouping 2 adjacent min terms are {(m_0, m_1), (m_1, m_3), (m_3, m_2), (m_2, m_0), (m_4, m_5), (m_5, m_7), (m_7, m_6), (m_6, m_4), (m_0, m_4), (m_1, m_5), (m_3, m_7) and (m_2, m_6)}.
- If x=0, then 3 variable K-map becomes 2 variable K-map.

4-Variable K-Map:

The number of cells in 4 variable K-map is sixteen, since the number of variables is four.





- There is only one possibility of grouping 16 adjacent min terms.
- Let R_1, R_2, R_3 and R_4 represents the min terms of first row, second row, third row and fourth row respectively. Similarly, C_1, C_2, C_3 and C_4 represents the min terms of first column, second column, third column and fourth column respectively. The possible

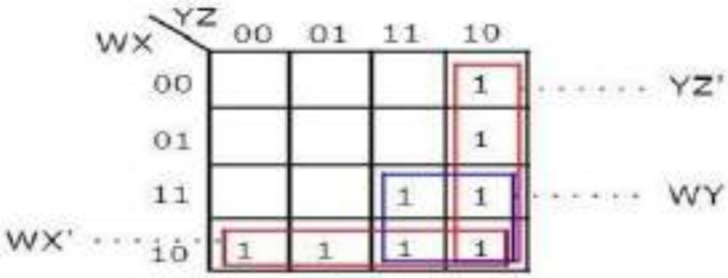
combinations of grouping 8 adjacent min terms are {(R_1, R_2), (R_2, R_3), (R_3, R_4), (R_4, R_1), (C_1, C_2), (C_2, C_3), (C_3, C_4), (C_4, C_1)}.

- If w=0, then 4 variable K-map becomes 3 variable K-map.

Example- **f(**W,X,Y,Z)**=** **∑m(2,6,8,9,10,11,14,15)**
using K-map.
The given Boolean function is in sum of products form. It is having 4 variables W, X, Y & Z. So, we require **4 variable K-map**.

WX \ YZ	00	01	11	10
00				1
01				1
11			1	1
10	1	1	1	1

The **4 variable K-map** with three **groupings** is

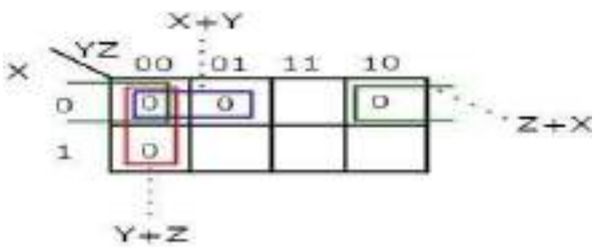


Therefore, the **simplified Boolean function** is **f = WX' + WY + YZ'**

Example- f(X,Y,Z)=∏M(0,1,2,4)
using K-map.
The given Boolean function is in product of Max terms form. It is having 3 variables X, Y & Z. So, we require 3 variable K-map.

X \ YZ	00	01	11	10
0	0	0		0
1	0			

The **3 variable K-map** with three **groupings** is



Therefore, the **simplified Boolean function** is

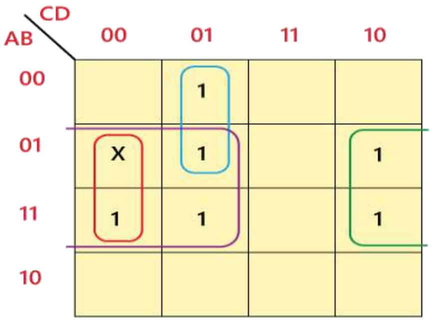
f = (X+Y)**. (**Y+Z)**. (**Z+X)

Don't care condition:

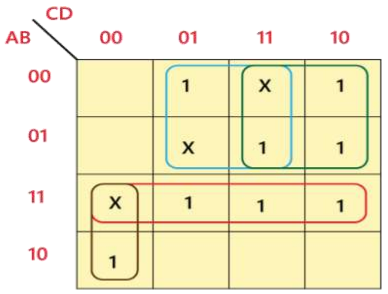
- The "Don't care" condition says that we can use the blank cells of a K-map to make a group of the variables.

- To make a group of cells, we can use the "don't care" cells as either 0 or 1, and if required, we can also ignore that cell.
- We mainly use the "don't care" cell to make a large group of cells.
- The cross(X) symbol is used to represent the "don't care" cell in K-map.
- This cross symbol represents an invalid combination.
- The "don't care" in excess-3 code are 0000, 0001, 0010, 1101, 1110, and 1111 because they are invalid combinations.
- Apart from this, the 4-bit BCD to Excess-3 code, the "don't care" are 1010, 1011, 1100, 1101, 1110, and 1111.

Example 1: Minimize f = ∑m(1,5,6,12,13,14) + d(4) in SOP minimal form



So, the minimized SOP form of the function is:
f = BC' + BD' + A'C'D **Example-2:**
Minimize the following function in SOP minimal form using K-Maps: F(A, B, C, D) = ∑m(1, 2, 6, 7, 8, 13, 14, 15) + d(3, 5, 12)



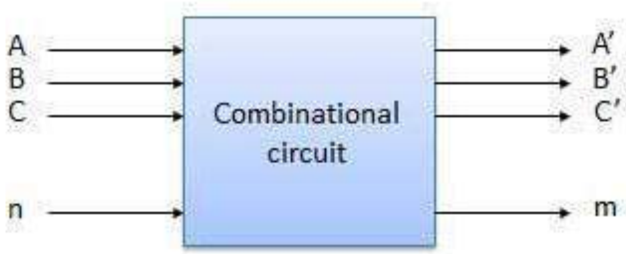
F = AC'D' + A'D + A'C + AB

UNIT-2Combinational logic circuit

Combinational circuit is a circuit in which we combine the different gates in the circuit, for example encoder, decoder, multiplexer and demultiplexer. Some of the characteristics of combinational circuits are following −

- The output of combinational circuit at any instant of time, depends only on the levels present at input terminals.
- The combinational circuit do not use any memory. The previous state of input does not have any effect on the present state of the circuit.
- A combinational circuit can have an n number of inputs and m number of outputs.

Block Diagram:



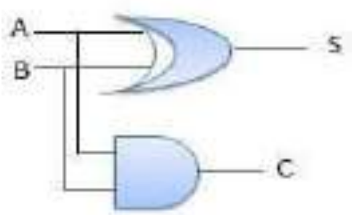
Half Adder

Half adder is a combinational logic circuit with two inputs and two outputs. The half adder circuit is designed to add two single bit binary number A and B. It is the basic building block for addition of two **single** bit numbers. This circuit has two outputs **carry** and **sum**. Block diagram

Truth Table

Inputs		Output	
A	B	S	C
0	0	0	0
0	1	1	0
1	0	1	0
1	1	0	1

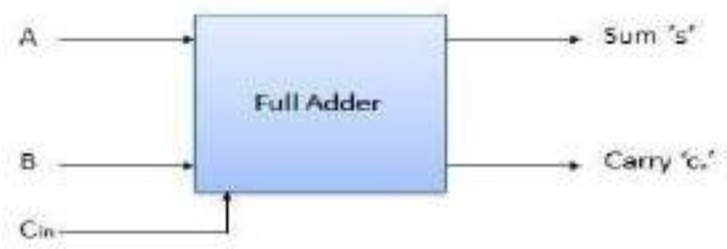
Circuit Diagram



Full Adder

Full adder is developed to overcome the drawback of Half Adder circuit. It can add two onebit numbers A and B, and carry c. The full adder is a three input and two output combinational circuit.

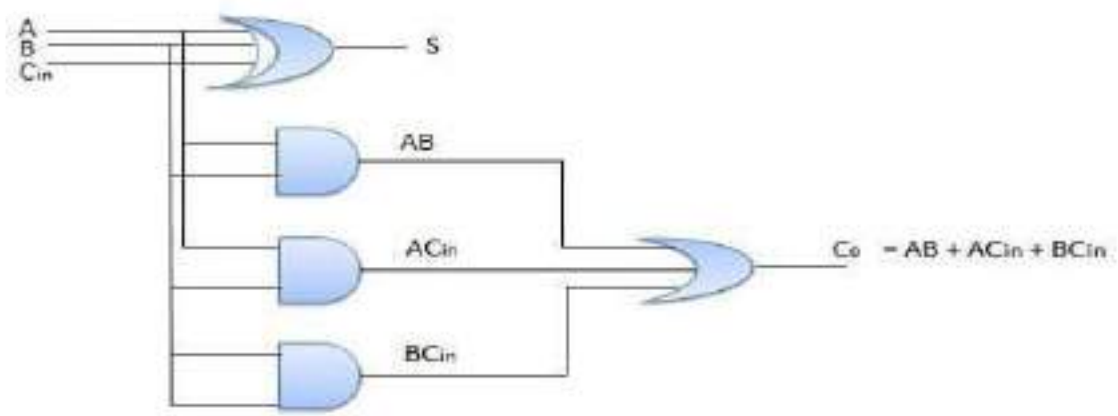
Block diagram



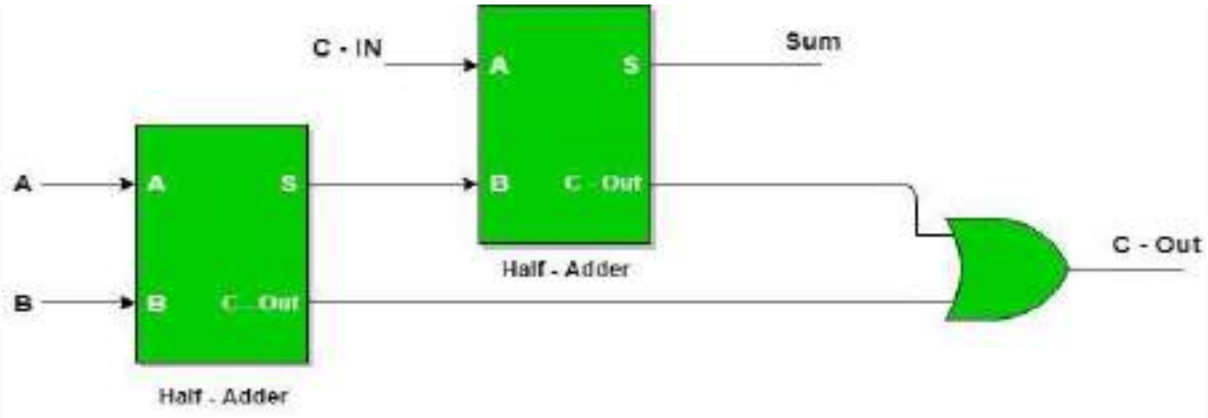
Truth Table

Inputs			Output	
A	B	C_{in}	S	C_o
0	0	0	0	0
0	0	1	1	0
0	1	0	1	0
0	1	1	0	1
1	0	0	1	0
1	0	1	0	1
1	1	0	0	1
1	1	1	1	1

Circuit Diagram

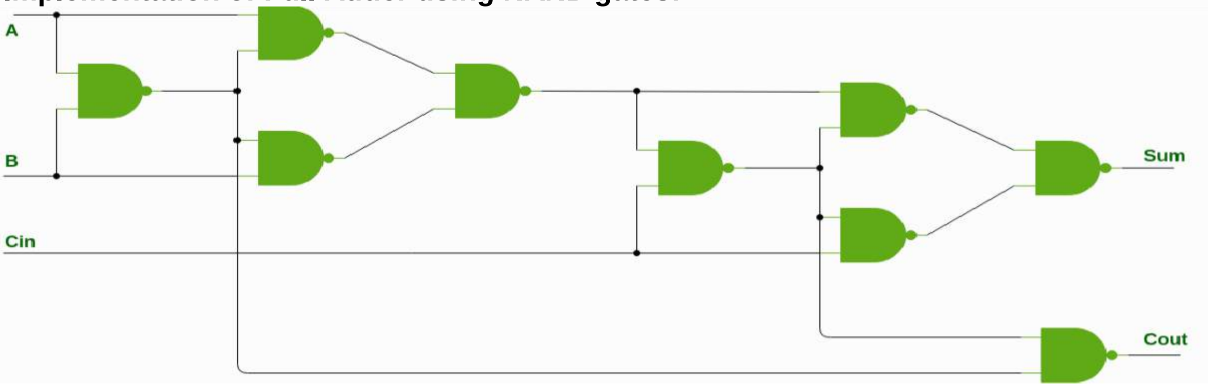


Implementation of Full Adder using Half Adders
2 Half Adders and a OR gate is required to implement a Full Adder.



With this logic circuit, two bits can be added together, taking a carry from the next lower order of magnitude, and sending a carry to the next higher order of magnitude.

Implementation of Full Adder using NAND gates:

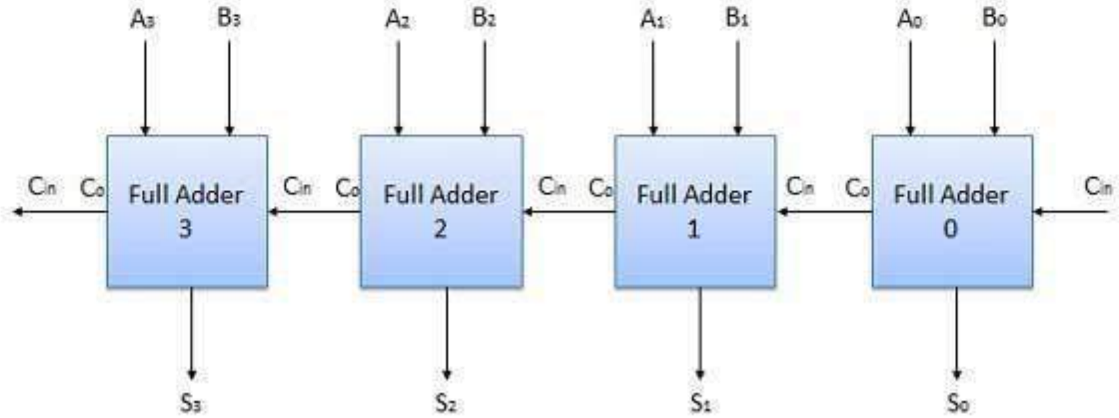


The Full Adder is capable of adding only two single digit binary number along with a carry input. But in practical we need to add binary numbers which are much longer than just one bit. To add two n-bit binary numbers we need to use the n-bit parallel adder. It uses a number of full adders in cascade. The carry output of the previous full adder is connected to carry input of the next full adder.

4 Bit Parallel Adder

In the block diagram, A_0 and B_0 represent the LSB of the four bit words A and B. Hence Full Adder-0 is the lowest stage. Hence its C_{in} has been permanently made 0. The rest of the connections are exactly same as those of n-bit parallel adder is shown in fig. The four-bit parallel adder is a very common logic circuit.

Block diagram



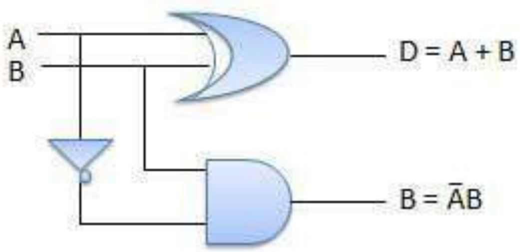
Half Subtractors

Half subtractor is a combination circuit with two inputs and two outputs (difference and borrow). It produces the difference between the two binary bits at the input and also produces an output (Borrow) to indicate if a 1 has been borrowed. In the subtraction (A-B), A is called as Minuend bit and B is called as Subtrahend bit.

Truth Table

Inputs		Output	
A	B	(A − B)	Borrow
0	0	0	0
0	1	1	1
1	0	1	0
1	1	0	0

Circuit Diagram



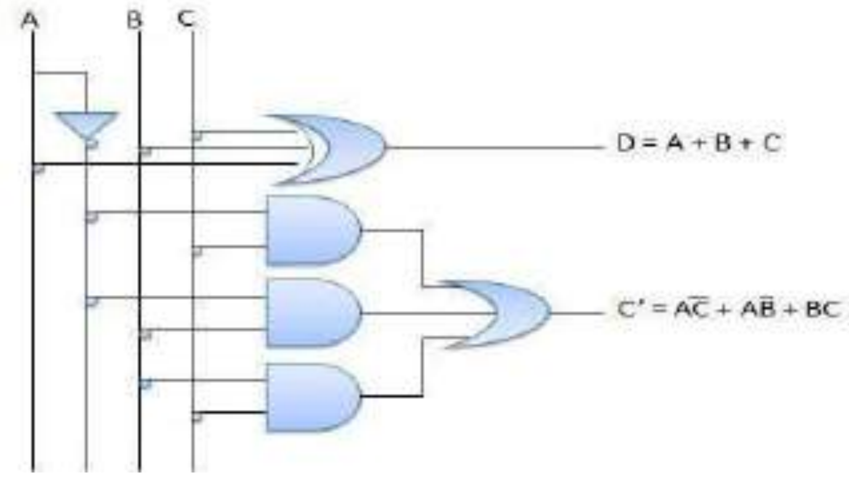
Full Subtractors

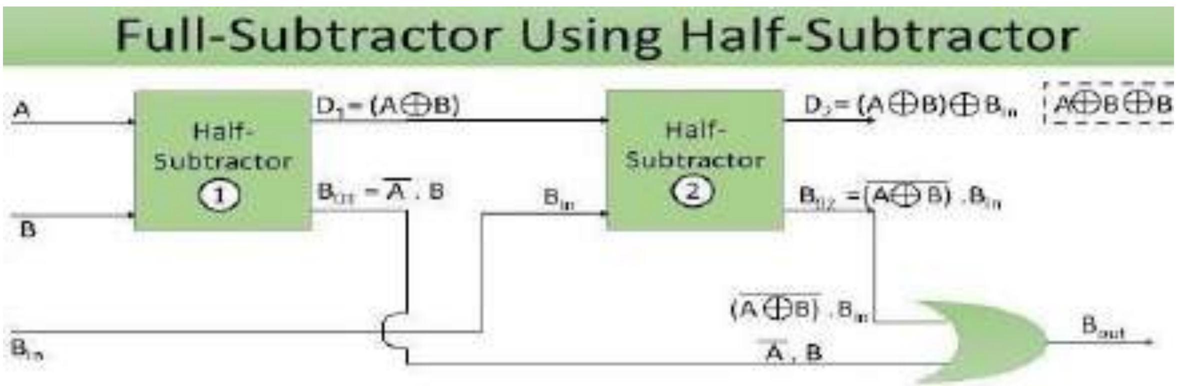
The disadvantage of a half subtractor is overcome by full subtractor. The full subtractor is a combinational circuit with three inputs A,B,C and two output D and C'. A is the 'minuend', B is 'subtrahend', C is the 'borrow' produced by the previous stage, D is the difference output and C' is the borrow output.

Truth Table

Inputs			Output	
A	B	C	(A-B-C)	C'
0	0	0	0	0
0	0	1	1	1
0	1	0	1	1
0	1	1	0	1
1	0	0	1	0
1	0	1	0	0
1	1	0	0	0
1	1	1	1	1

Circuit Diagram

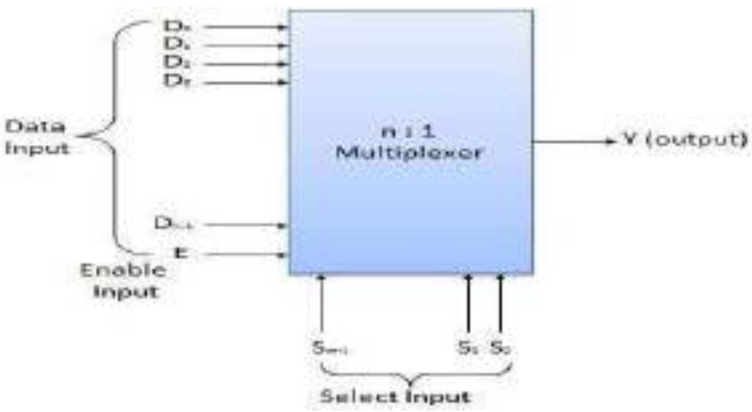




Multiplexers

Multiplexer is a special type of combinational circuit. There are n-data inputs, one output and m select inputs with 2m = n. It is a digital circuit which selects one of the n data inputs and routes it to the output. The selection of one of the n inputs is done by the selected inputs. Depending on the digital code applied at the selected inputs, one out of n data sources is selected and transmitted to the single output Y. E is called the strobe or enable input which is useful for the cascading. It is generally an active low terminal that means it will perform the required operation when it is low.

Block diagram



Multiplexers come in multiple variations

- 2 : 1 multiplexer
- 4 : 1 multiplexer
- 16 : 1 multiplexer
- 32 : 1 multiplexer

Block Diagram

Truth Table

Enable	Select	Output
E	S	Y
0	x	0
1	0	D_0
1	1	D_1

x = Don't care

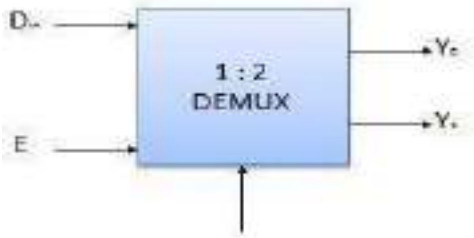
Demultiplexers

A demultiplexer performs the reverse operation of a multiplexer i.e. it receives one input and distributes it over several outputs. It has only one input, n outputs, m select input. At a time only one output line is selected by the select lines and the input is transmitted to the selected output line. A de-multiplexer is equivalent to a single pole multiple way switch as shown in fig.

Demultiplexers comes in multiple variations.

- 1 : 2 demultiplexer
- 1 : 4 demultiplexer
- 1 : 16 demultiplexer
- 1 : 32 demultiplexer

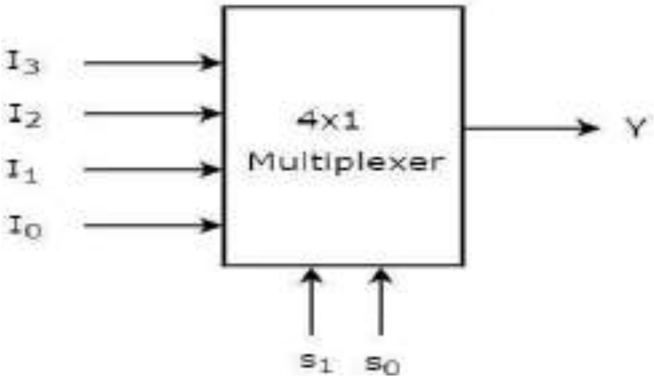
Block diagram



Truth Table

4:1 Multiplexer

4:1 Multiplexer has four data inputs I_3, I_2, I_1& I_0, two selection lines s_1& s_0 and one output Y. The **block diagram** of 4x1 Multiplexer is shown in the following figure.



One of these 4 inputs will be connected to the output based on the combination of inputs present at these two selection lines.

Truth table of 4:1 Multiplexer is shown below.

Selection Lines		Output
S_1	S_0	Y
0	0	I_0

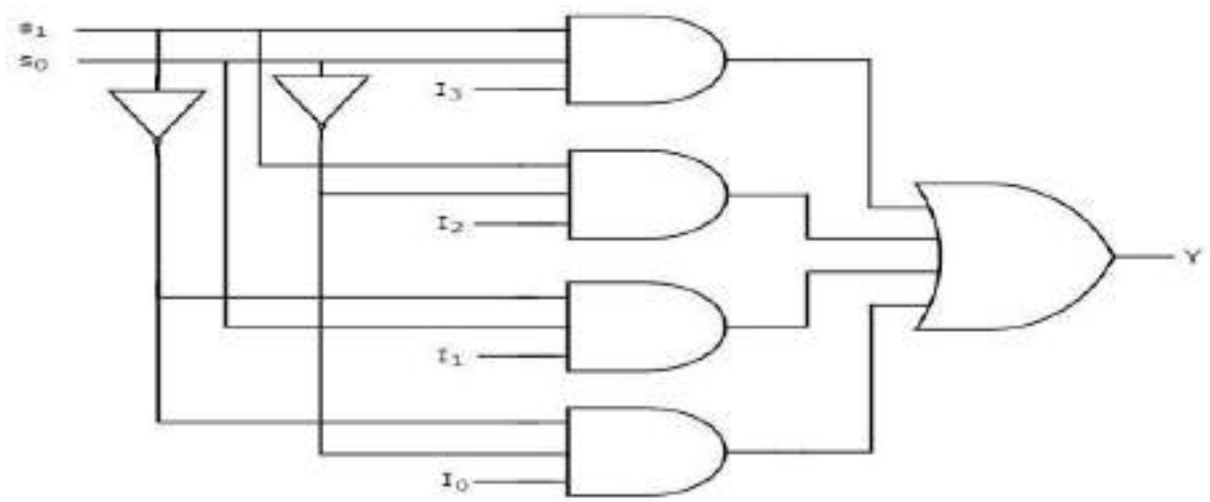
0	1	I_1
1	0	I_2
1	1	I_3

From Truth table, we can directly write the **Boolean function** for output, Y as

$$Y = S1'S0'I_0 + S1'S0I_1 + S1S0'I_2 + S1S0I_3$$

We can implement this Boolean function using Inverters, AND gates & OR gate.

The **circuit diagram** of 4:1 multiplexer is shown in the following figure.



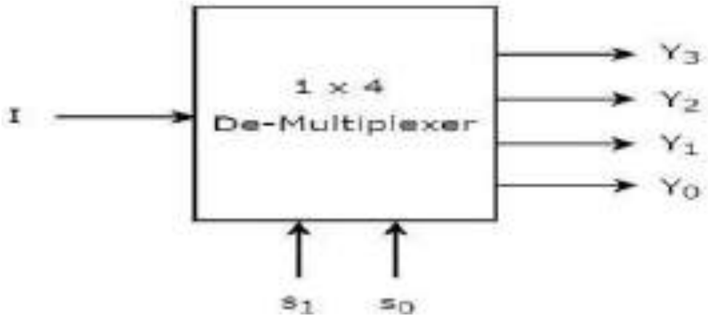
We can easily understand the operation of the above circuit. Similarly, you can implement 8x1 Multiplexer and 16x1 multiplexer by following the same procedure.

Enable	Select	Output	
E	S	Y0	Y1
0	x	0	0
1	0	0	D_{in}
1	1	D_{in}	0

x = Don't care

1:4 De-Multiplexer

1:4 De-Multiplexer has one input I, two selection lines, s_1& s_0 and four outputs Y_3, Y_2, Y_1&Y_0. The **block diagram** of 1:4 De-Multiplexer is shown in the following figure.



The single input ‘I’ will be connected to one of the four outputs, Y_3 to Y_0 based on the values of selection lines s_1& s0. The **Truth table** of 1x4 De-Multiplexer is shown below.

Selection Inputs		**Outputs**			
S_1	**S_0**	**Y_3**	**Y_2**	**Y_1**	**Y_0**
0	0	0	0	0	**I**
0	1	0	0	**I**	0
1	0	0	**I**	0	0
1	1	**I**	0	0	0

From the above Truth table, we can directly write the **Boolean functions** for each output as

$$Y3 = s_1 s_0 I$$

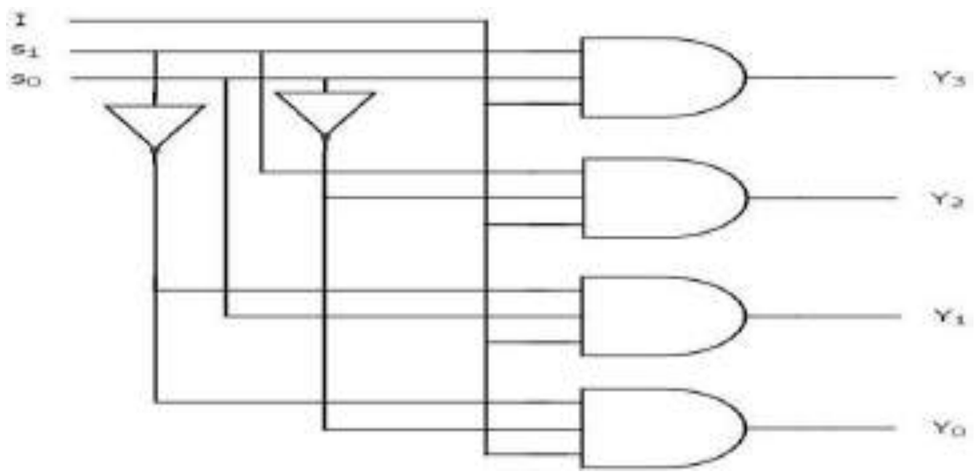
$$Y2 = s1s0'I$$

$$Y1 = s1's0I$$

$$Y0 = s1's0'I$$

We can implement these Boolean functions using Inverters & 3-input AND gates.

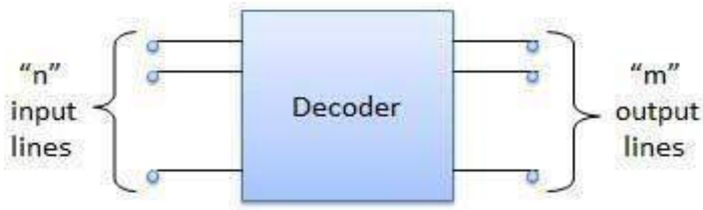
The **circuit diagram** of 1:4 De-Multiplexer is shown in the following figure.



Decoder

A decoder is a combinational circuit. It has n input and to a maximum m = 2n outputs. Decoder is identical to a demultiplexer without any data input. It performs operations which are exactly opposite to those of an encoder.

Block diagram



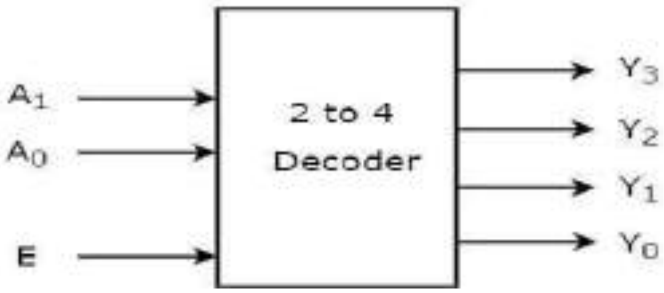
Examples of Decoders are following.

- Code converters
- BCD to seven segment decoders

2 to 4 Decoder

Let 2 to 4 Decoder has two inputs A_1& A_0 and four outputs Y_3, Y_2, Y_1& Y_0.

The **block diagram** of 2 to 4 decoder is shown in the following figure.



One of these four outputs will be '1' for each combination of inputs when enable, E is '1'.

The **Truth table** of 2 to 4 decoder is shown below.

Enable	Inputs	Outputs

E	A_1	A_0	Y_3	Y_2	Y_1	Y_0
0	x	x	0	0	0	0
1	0	0	0	0	0	1
1	0	1	0	0	1	0
1	1	0	0	1	0	0
1	1	1	1	0	0	0

From Truth table, we can write the **Boolean functions** for each output as

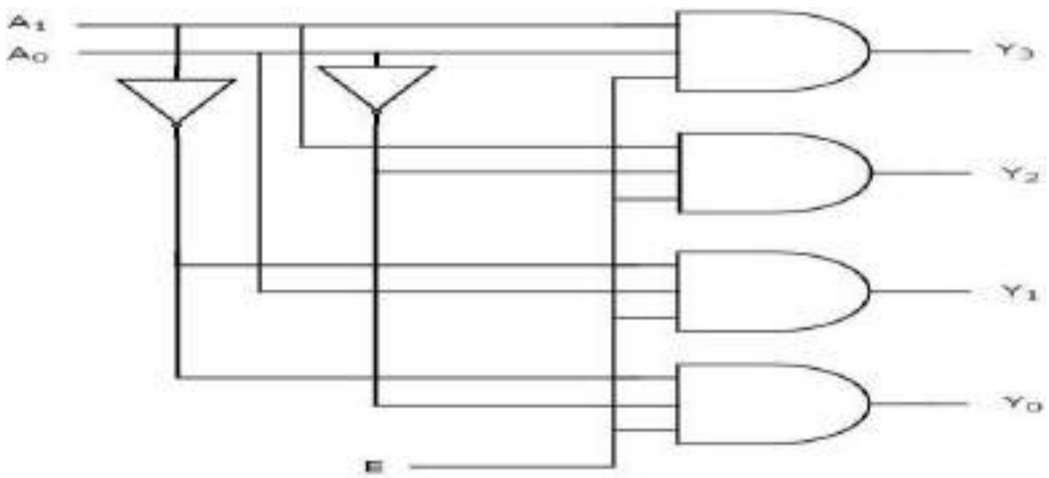
$$Y3 = E.A1.A0$$

$$Y2 = E.A1.A0'$$

$$Y1 = E.A1'.A0$$

$$Y0 = E.A1'.A0'$$

Each output is having one product term. So, there are four product terms in total. We can implement these four product terms by using four AND gates having three inputs each & two inverters. The **circuit diagram** of 2 to 4 decoder is shown in the following figure.

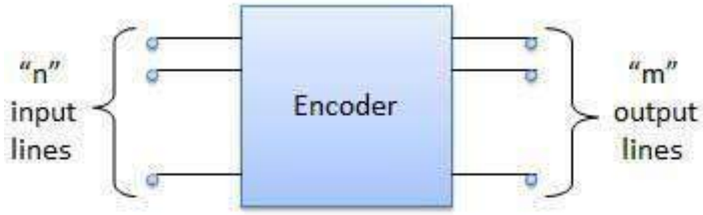


Therefore, the outputs of 2 to 4 decoder are nothing but the **min terms** of two input variables A_1& A_0, when enable, E is equal to one. If enable, E is zero, then all the outputs of decoder will be equal to zero.

Similarly, 3 to 8 decoder produces eight min terms of three input variables A_2, A_1& A_0 and 4 to 16 decoder produces sixteen min terms of four input variables A_3, A_2, A_1& A_0.

Encoder

Encoder is a combinational circuit which is designed to perform the inverse operation of the decoder. An encoder has n number of input lines and m number of output lines. An encoder produces an m bit binary code corresponding to the digital input number. The encoder accepts an n input digital word and converts it into an m bit another digital word. Block diagram



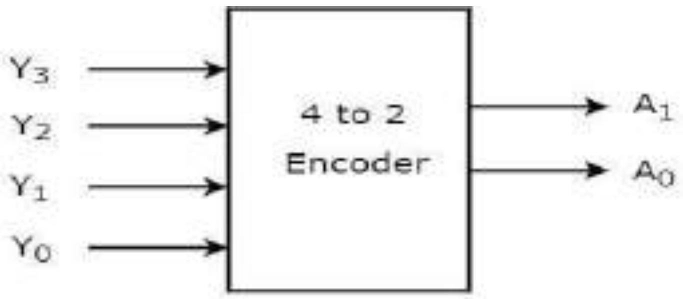
Examples of Encoders are following.

- Priority encoders
- Decimal to BCD encoder
- Octal to binary encoder
- Hexadecimal to binary encoder

4 to 2 Encoder

Let 4 to 2 Encoder has four inputs Y_3, Y_2, Y_1& Y_0 and two outputs A_1& A_0.

The **block diagram** of 4 to 2 Encoder is shown in the following figure.



At any time, only one of these 4 inputs can be '1' in order to get the respective binary code at the output. The **Truth table** of 4 to 2 encoder is shown below.

Inputs				**Outputs**	
Y_3	Y_2	Y_1	Y_0	A_1	A_0
0	0	0	1	0	0

0	0	1	0	0	1
0	1	0	0	1	0
1	0	0	0	1	1

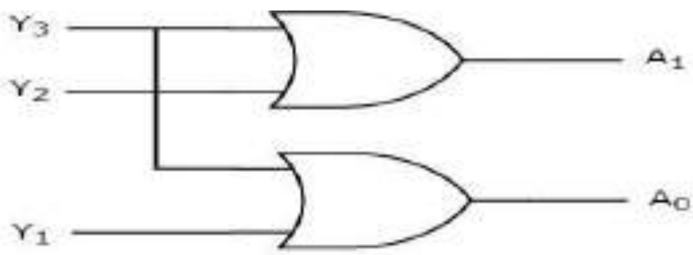
From Truth table, we can write the **Boolean functions** for each output as

$$A1 = Y3 + Y2$$

$$A0 = Y3 + Y1$$

We can implement the above two Boolean functions by using two input OR gates.

The **circuit diagram** of 4 to 2 encoder is shown in the following figure.

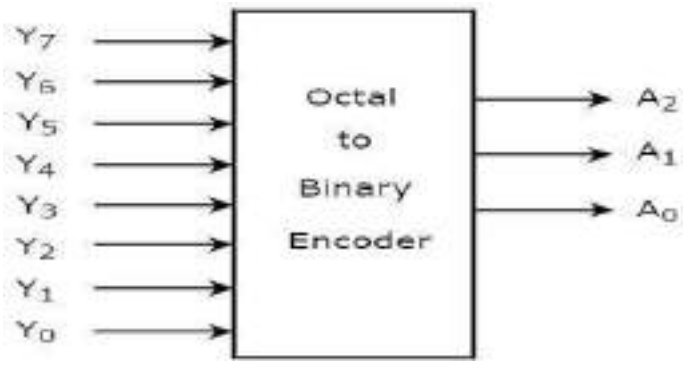


The above circuit diagram contains two OR gates. These OR gates encode the four inputs with two bits

Octal to Binary Encoder

Octal to binary Encoder has eight inputs, Y_7 to Y_0 and three outputs A_2, A_1& A_0. Octal to binary encoder is nothing but 8 to 3 encoder.

The **block diagram** of octal to binary Encoder is shown in the following figure.



At any time, only one of these eight inputs can be '1' in order to get the respective binary code. The **Truth table** of octal to binary encoder is shown below.

Inputs	Outputs

Y_7	Y_6	Y_5	Y_4	Y_3	Y_2	Y_1	Y_0	A_2	A_1	A_0
0	0	0	0	0	0	0	1	0	0	0
0	0	0	0	0	0	1	0	0	0	1
0	0	0	0	0	1	0	0	0	1	0
0	0	0	0	1	0	0	0	0	1	1
0	0	0	1	0	0	0	0	1	0	0
0	0	1	0	0	0	0	0	1	0	1
0	1	0	0	0	0	0	0	1	1	0
1	0	0	0	0	0	0	0	1	1	1

From Truth table, we can write the **Boolean functions** for each output as

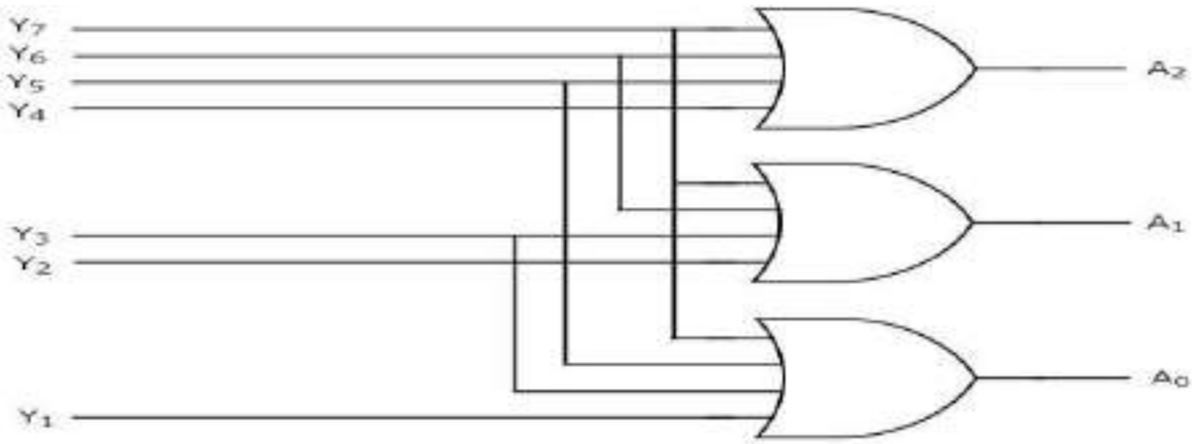
$$A2 = Y7 + Y6 + Y5 + Y4$$

$$A1 = Y7 + Y6 + Y3 + Y2$$

$$A0 = Y7 + Y5 + Y3 + Y1$$

We can implement the above Boolean functions by using four input OR gates.

The **circuit diagram** of octal to binary encoder is shown in the following figure.

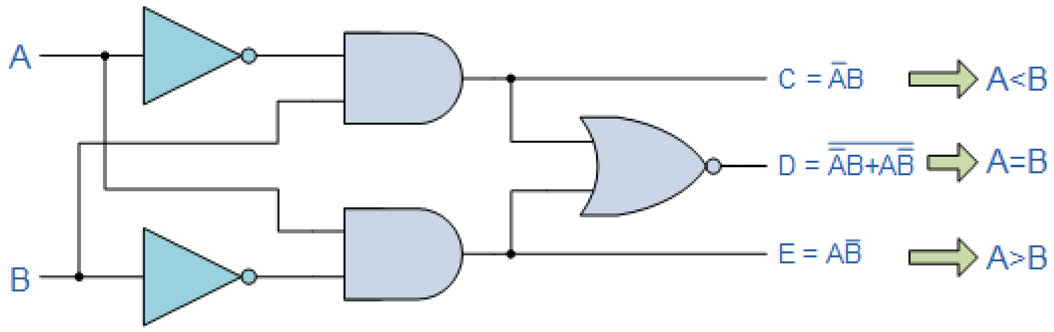


The above circuit diagram contains three 4-input OR gates. These OR gates encode the eight inputs with three bits.

Digital comparator

The Digital Comparator is another very useful combinational logic circuit used to compare the value of two binary digits.

1-bit Digital Comparator Circuit



Then the operation of a 1-bit digital comparator is given in the following Truth Table. **Digital Comparator Truth Table**

Inputs		Outputs		
B	A	A > B	A = B	A < B
0	0	0	1	0

0	1	1	0	0
1	0	0	0	1
1	1	0	1	0

2-bit Magnitude Comparator

A comparator that compares two binary numbers (each number having 2 bits) and produces three outputs based on the relative magnitudes of given binary bits is called a 2-bit magnitude comparator.

Truth Table

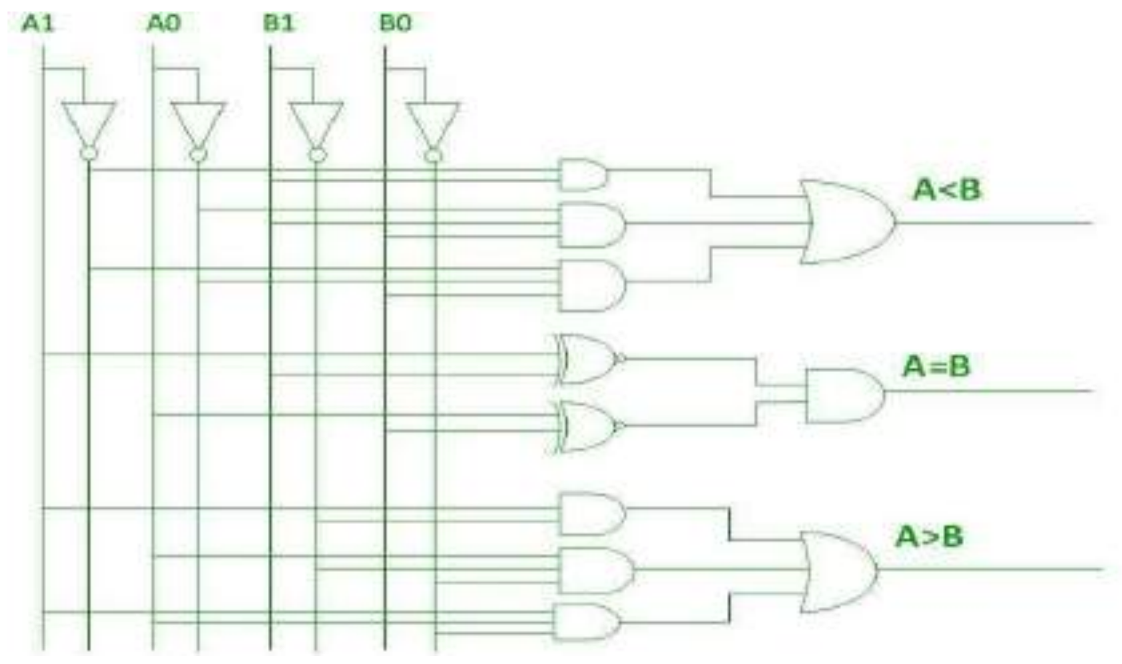
A1	A0	B1	B0	A<B	A=B	A>B
0	0	0	0	0	1	0
0	0	0	1	1	0	0
0	0	1	0	1	0	0
0	0	1	1	1	0	0
0	1	0	0	0	0	1
0	1	0	1	0	1	0
0	1	1	0	1	0	0
0	1	1	1	1	0	0
1	0	0	0	0	0	1
1	0	0	1	0	0	1

1	0	1	0	0	1	0
1	0	1	1	1	0	0
1	1	0	0	0	0	1
1	1	0	1	0	0	1
1	1	1	0	0	0	1
1	1	1	1	0	1	0

The truth table derives the expressions of A<B, A>B, and A=B as below

$$\boldsymbol{A < B - A1'B1' + A0'B1B0 + A1'A0'B0}$$
$$\boldsymbol{A > B - A1B1' + A0B1'B0' + A1A0B0'}$$
$$\boldsymbol{A = B - (A0\ Ex - Nor\ B0)\ (A1\ Ex - Nor\ B1)}$$

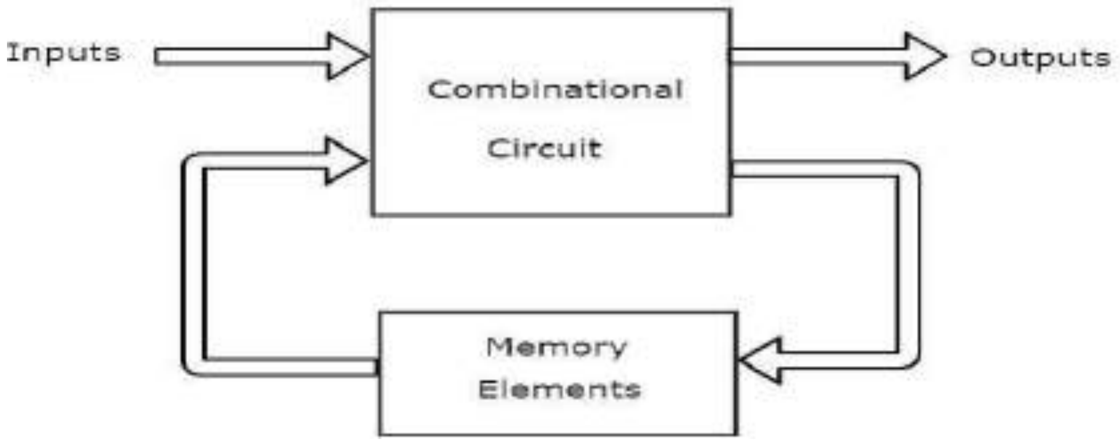
With these expressions, the Circuit diagram can be as follows



UNIT-3

Sequential Logic Circuits

Sequential circuit contains a set of inputs and outputs. The outputs of sequential circuit depend not only on the combination of present inputs but also on the previous outputs. Previous output is nothing but the **present state**. Therefore, sequential circuits contain combinational circuits along with memory storage elements. Some sequential circuits may not contain combinational circuits, but only memory elements.



Following table shows the **differences** between combinational circuits and sequential circuits.

Combinational Circuits	Sequential Circuits
Outputs depend only on present inputs.	Outputs depend on both present inputs and present state.
Feedback path is not present.	Feedback path is present.
Memory elements are not required.	Memory elements are required.
Clock signal is not required.	Clock signal is required.
Easy to design.	Difficult to design.

Types of Sequential Circuits

Following are the two types of sequential circuits −

- Asynchronous sequential circuits
- Synchronous sequential circuits

Asynchronous sequential circuits

If some or all the outputs of a sequential circuit do not change affect with respect to active transition of clock signal, then that sequential circuit is called as **Asynchronous sequential circuit**. That means, all the outputs of asynchronous sequential circuits do not change affect at the same time. Therefore, most of the outputs of asynchronous sequential circuits are **not in synchronous** with either only positive edges or only negative edges of clock signal.

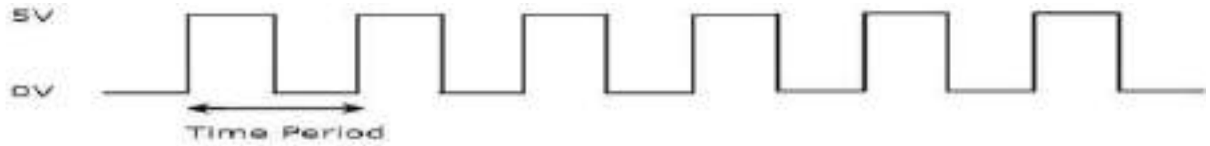
Synchronous sequential circuits

If all the outputs of a sequential circuit change affect with respect to active transition of clock signal, then that sequential circuit is called as **Synchronous sequential circuit**. That means, all the outputs of synchronous sequential circuits change affect at the same time. Therefore, the outputs of synchronous sequential circuits are in synchronous with either only positive edges or only negative edges of clock signal.

Clock Signal and Triggering

Clock signal

Clock signal is a periodic signal and its ON time and OFF time need not be the same. We can represent the clock signal as a **square wave**, when both its ON time and OFF time are same. This clock signal is shown in the following figure.



Types of Triggering

Following are the two possible types of triggering that are used in sequential circuits.

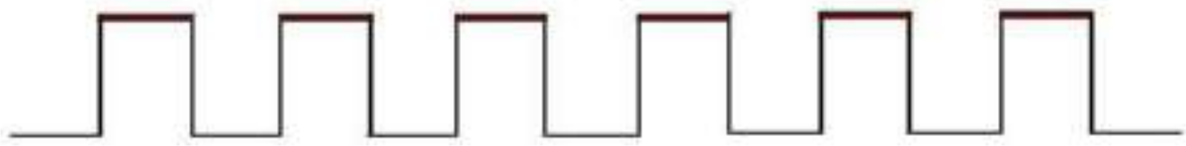
- Level triggering
- Edge triggering

Level triggering

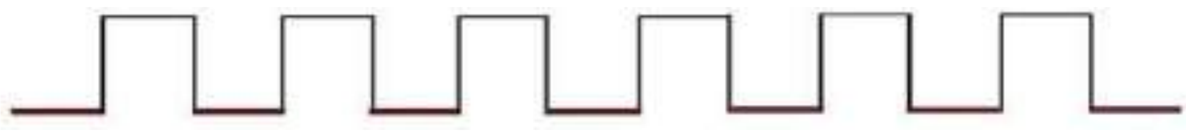
There are two levels, namely logic High and logic Low in clock signal. Following are the two **types of level triggering**.

- Positive level triggering
- Negative level triggering

If the sequential circuit is operated with the clock signal when it is in **Logic High**, then that type of triggering is known as **Positive level triggering**. It is highlighted in below figure.



If the sequential circuit is operated with the clock signal when it is in **Logic Low**, then that type of triggering is known as **Negative level triggering**. It is highlighted in the following figure.



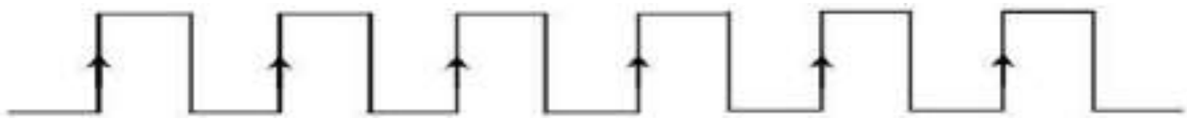
Edge triggering

There are two types of transitions that occur in clock signal. That means, the clock signal transitions either from Logic Low to Logic High or Logic High to Logic Low.

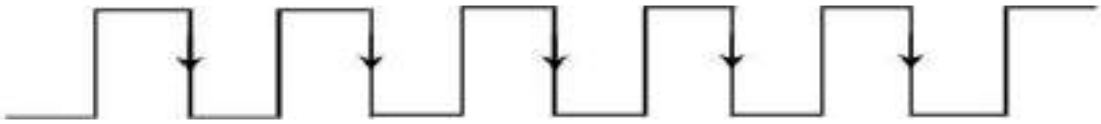
Following are the two **types of edge triggering** based on the transitions of clock signal.

- Positive edge triggering
- Negative edge triggering

If the sequential circuit is operated with the clock signal that is transitioning from Logic Low to Logic High, then that type of triggering is known as **Positive edge triggering**. It is also called as rising edge triggering. It is shown in the following figure.



If the sequential circuit is operated with the clock signal that is transitioning from Logic High to Logic Low, then that type of triggering is known as **Negative edge triggering**. It is also called as falling edge triggering. It is shown in the following figure.



There are two types of memory elements based on the type of triggering that is suitable to operate it.

- Latches
- Flip-flops

Latches operate with enable signal, which is **level sensitive**. Whereas, flip-flops are edge sensitive.

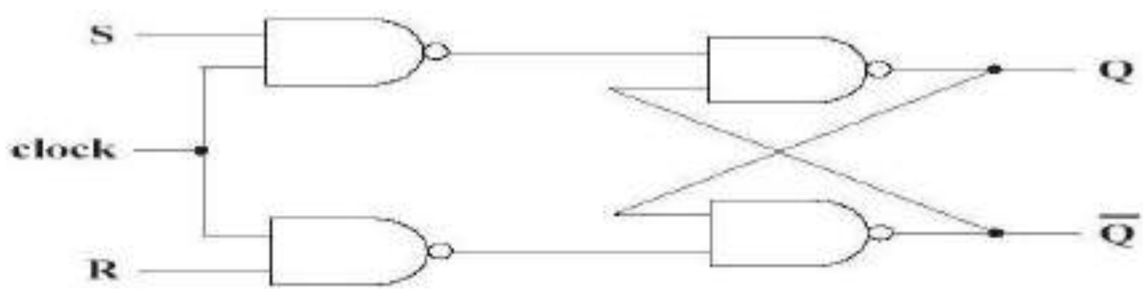
There are 4 types of flip flops:

- SR Flip-Flop
- D Flip-Flop
- JK Flip-Flop
- T Flip-Flop

SR Flip-Flop

SR flip-flop operates with only positive clock transitions or negative clock transitions. Whereas, SR latch operates with enable signal.

The **circuit diagram** of SR flip-flop is shown in the following figure.



This circuit has two inputs S & R and two outputs Q& Q'. The operation of SR flipflop is similar to SR Latch. But, this flip-flop affects the outputs only when positive transition of the clock signal is applied instead of active enable.

The following table shows the **state table** of SR flip-flop.

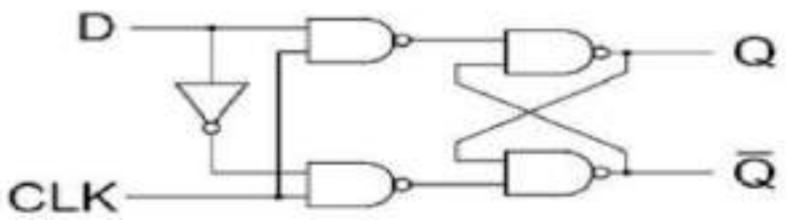
S	**R**	**Q**t+1
0	0	Q
0	1	0
1	0	1
1	1	-

Here, Q & Qt+1 are present state & next state respectively. So, SR flip-flop can be used for one of these three functions such as Hold, Reset & Set based on the input conditions, when positive transition of clock signal is applied.

D Flip-Flop

D flip-flop operates with only positive clock transitions or negative clock transitions. Whereas, D latch operates with enable signal. That means, the output of D flip-flop is insensitive to the changes in the input, D except for active transition of the clock signal.

The **circuit diagram** of D flip-flop is shown in the following figure.



This circuit has single input D and two outputs Q & Q'. The operation of D flip-flop is similar to D Latch. But, this flip-flop affects the outputs only when positive transition of the clock signal is applied instead of active enable.

The following table shows the **state table** of D flip-flop.

D	**Qt + 1**
0	0
1	1

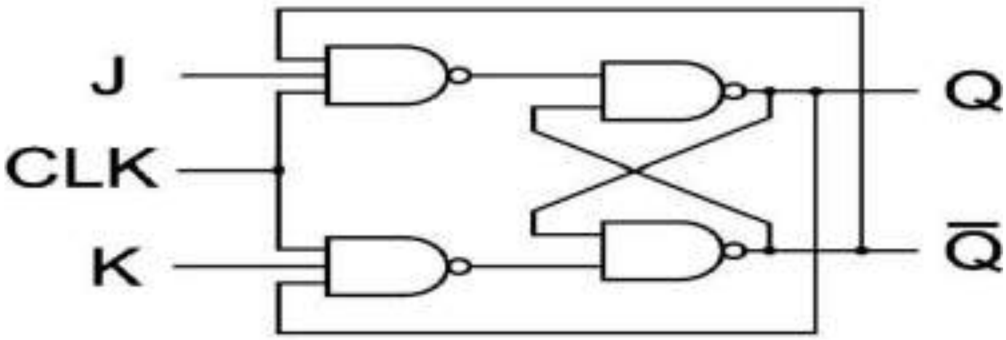
Therefore, D flip-flop always Hold the information, which is available on data input, D of earlier positive transition of clock signal.

D flip-flops can be used in registers, **shift registers** and some of the counters.

JK Flip-Flop

JK flip-flop is the modified version of SR flip-flop. It operates with only positive clock transitions or negative clock transitions.

The **circuit diagram** of JK flip-flop is shown in the following figure.



This circuit has two inputs J & K and two outputs Q & Q'. The operation of JK flip-flop is similar to SR flip-flop.

The following table shows the **state table** of JK flip-flop.

J	**K**	**Q**t+1
0	0	Q
0	1	0
1	0	1
1	1	Q'

Here, Q & Qt+1 are present state & next state respectively. So, JK flip-flop can be used for one of these four functions such as Hold, Reset, Set & Complement of present state based on the input conditions, when positive transition of clock signal is applied.

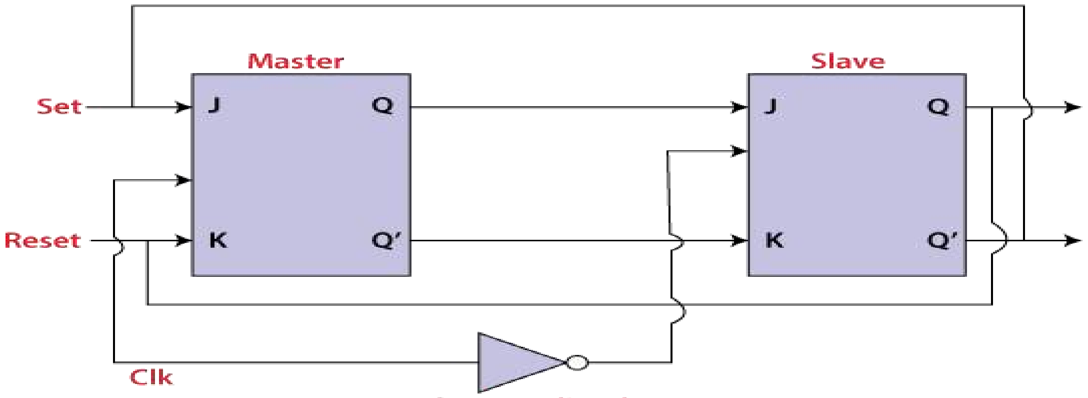
Master-Slave JK Flip Flop

In "JK Flip Flop", when both the inputs and CLK set to 1 for a long time, then Q output toggle until the CLK is 1. Thus, the uncertain or unreliable output produces. This problem is referred to as a **race-round condition** in JK flip-flop and avoided by ensuring that the CLK set to 1 only for a very short time.

Explanation

The master-slave flip flop is constructed by combining two J K flip flop. These flip flops are connected in a series configuration. In these two flip flops, the 1st flip flop work as "master", called the master flip flop, and the 2nd work as a "slave", called slave flip flop.

In "master-slave flip flop", apart from these two flip flops, an inverter or NOT gate is also used. For passing the inverted clock pulse to the "slave" flip flop, the inverter is connected to the clock's pulse. In simple words, when CP set to false for "master", then CP is set to true for "slave", and when CP set to true for "master", then CP is set to false for "slave".

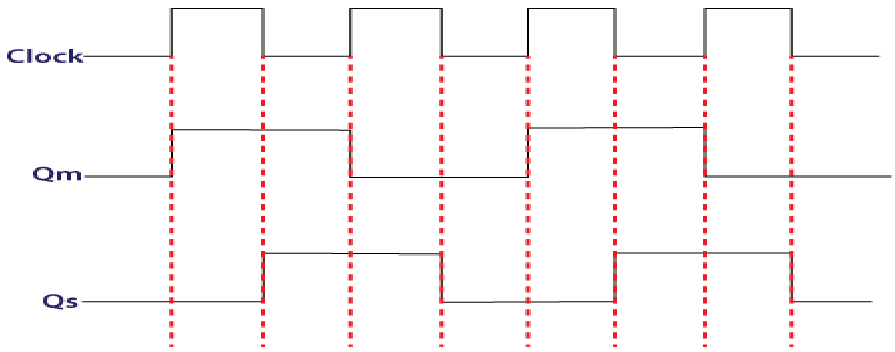


Master-slave JK Flip-Flop

Working:

- When the clock pulse is true, the slave flip flop will be in the isolated state, and the system's state may be affected by the J and K inputs. The "slave" remains isolated until the CP is 1. When the CP set to 0, the master flip-flop passes the information to the slave flip flop to obtain the output.
- The master flip flop responds first from the slave because the master flip flop is the positive level trigger, and the slave flip flop is the negative level trigger.
- The output Q'=1 of the master flip flop is passed to the slave flip flop as an input K when the input J set to 0 and K set to 1. The clock forces the slave flip flop to work as reset, and then the slave copies the master flip flop.
- When J=1, and K=0, the output Q=1 is passed to the J input of the slave. The clock's negative transition sets the slave and copies the master.
- The master flip flop toggles on the clock's positive transition when the inputs J and K set to 1. At that time, the slave flip flop toggles on the clock's negative transition.
- The flip flop will be disabled, and Q remains unchanged when both the inputs of the JK flip flop set to 0.

Timing Diagram of a Master Flip Flop:

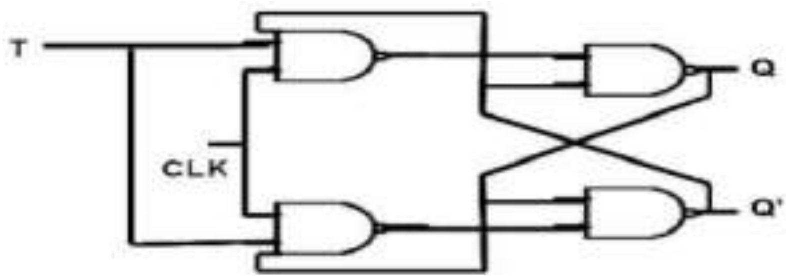


- When the clock pulse set to 1, the output of the master flip flop will be one until the clock input remains 0.
- When the clock pulse becomes high again, then the master's output is 0, which will be set to 1 when the clock becomes one again.
- The master flip flop is operational when the clock pulse is 1. The slave's output remains 0 until the clock is not set to 0 because the slave flip flop is not operational.
- The slave flip flop is operational when the clock pulse is 0. The output of the master remains one until the clock is not set to 0 again.
- Toggling occurs during the entire process because the output changes once in the cycle.

T Flip-Flop

T flip-flop is the simplified version of JK flip-flop. It is obtained by connecting the same input 'T' to both inputs of JK flip-flop. It operates with only positive clock transitions or negative clock transitions.

The **circuit diagram** of T flip-flop is shown in the following figure.



This circuit has single input T and two outputs Q & Q'. The operation of T flip-flop is same as that of JK flip-flop. Here, we considered the inputs of JK flip-flop as **J = T** and **K = T** in order to utilize the modified JK flip-flop for 2 combinations of inputs. So, we eliminated the other two combinations of J & K, for which those two values are complement to each other in T flipflop.

The following table shows the **state table** of T flip-flop.

D	**Q**t+1

0	Q
1	Q’

Here, Q & Qt+1 are present state & next state respectively. So, T flip-flop can be used for one of these two functions such as Hold, & Complement of present state based on the input conditions, when positive transition of clock signal is applied.
The output of T flip-flop always toggles for every positive transition of the clock signal, when input T remains at logic High 11. Hence, T flip-flop can be used in **counters**.

UNIT-4

Registers, memories and PLD

Shift register

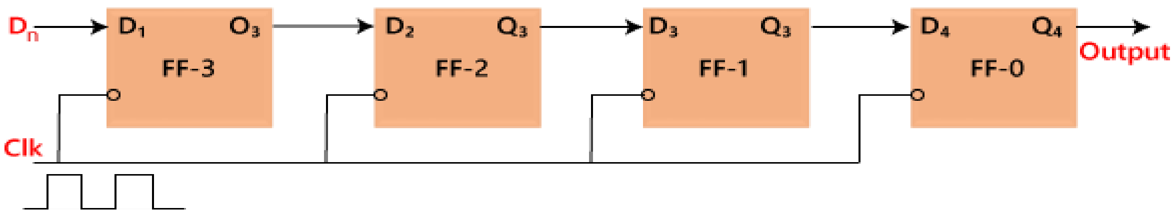
- Flip flops can be used to store a single bit of binary data (1or 0).
- However, in order to store multiple bits of data, we need multiple flip flops. N flip flops are to be connected in an order to store n bits of data.
- A **Register** is a device which is used to store such information. It is a group of flip flops connected in series used to store multiple bits of data.
- The information stored within these registers can be transferred with the help of **shift registers**.
- Shift Register is a group of flip flops used to store multiple bits of data. The bits stored in such registers can be made to move within the registers and in/out of the registers by applying clock pulses.
- An n-bit shift register can be formed by connecting n flip-flops where each flip flop stores a single bit of data.

Shift registers are basically of 4 types. These are:

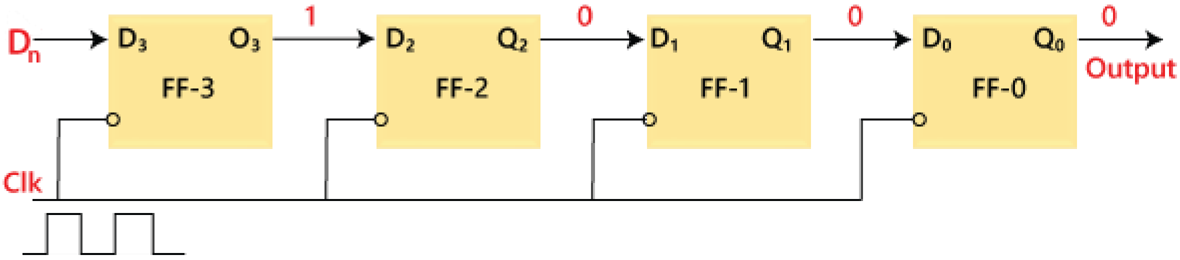
1. Serial In Serial Out shift register
2. Serial In parallel Out shift register
3. Parallel In Serial Out shift register
4. Parallel In parallel Out shift register

Serial-In Serial-Out Shift Register (SISO) –

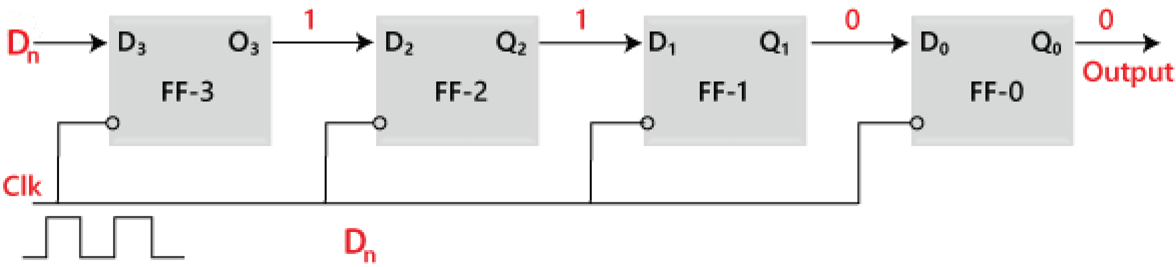
The shift register, which allows serial input (one bit after the other through a single data line) and produces a serial output is known as Serial-In Serial-Out shift register. Since there is only one output, the data leaves the shift register one bit at a time in a serial pattern, thus the name Serial-In Serial-Out Shift Register. The circuit consists of four D flip-flops which are connected in a serial manner. All these flip-flops are synchronous with each other since the same clock signal is applied to each flip flop. The main use of a SISO is to act as a delay element.



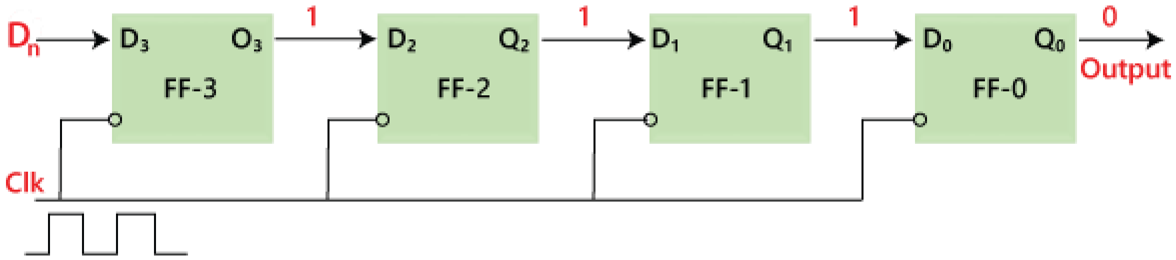
D_n
D_1
O_3
FF-3
D_2
Q_3
FF-2
D_3
Q_3
FF-1
D_4
Q_4
FF-0
Output
Clk

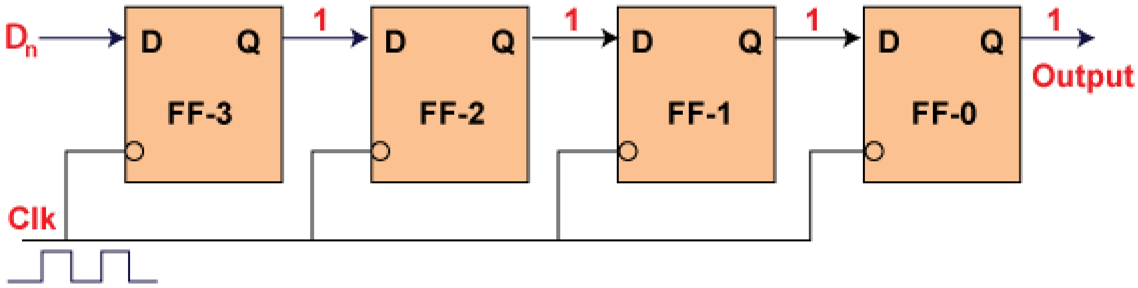


D_n
D_3
O_3
FF-3
1
D_2
Q_2
FF-2
0
D_1
Q_1
FF-1
0
D_0
Q_0
FF-0
0
Output
Clk



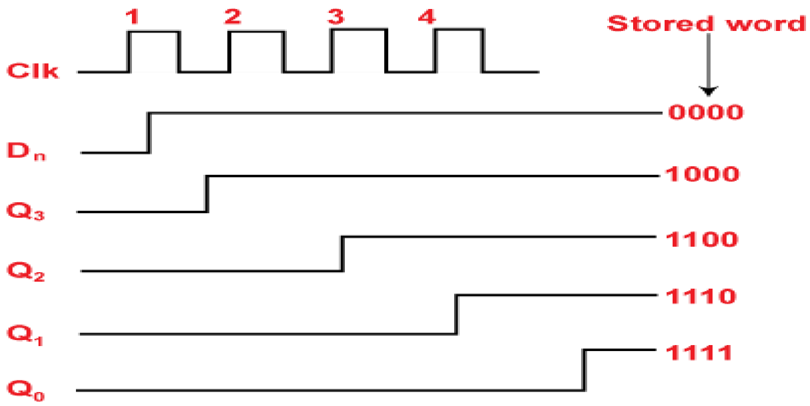
D_n
D_3
O_3
FF-3
1
D_2
Q_2
FF-2
1
D_1
Q_1
FF-1
0
D_0
Q_0
FF-0
0
Output
Clk
D_n





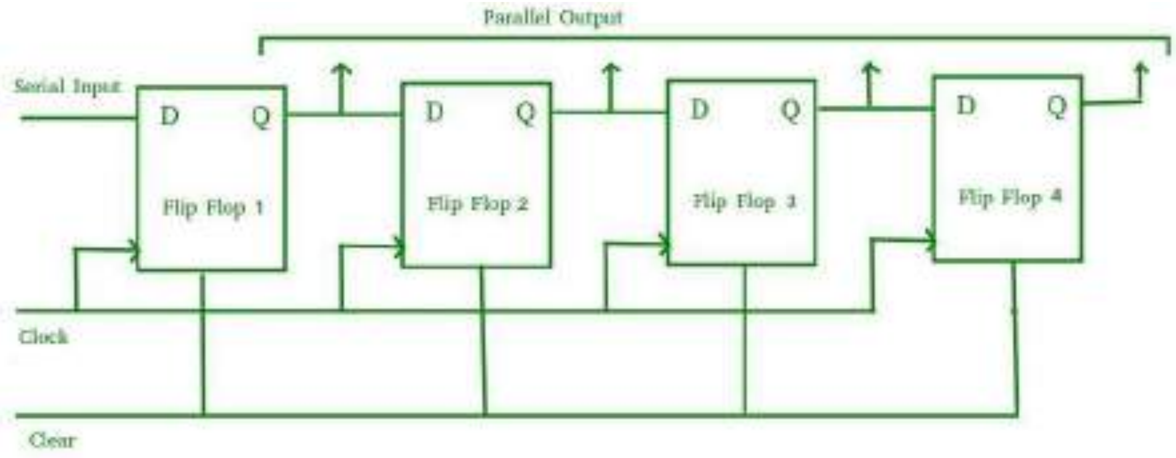
	Clk	$D_n = Q_3$	$Q_3 = D_2$	$Q_2 = D_1$	$Q_1 = D_0$	Q_0
Initially			0	0	0	0
(1)	↓	1	1	0	0	0
(2)	↓	1	1	1	0	0
(3)	↓	1	1	1	1	0
(4)	↓	1	1	1	1	1

Direction of data travel



Serial-In Parallel-Out shift Register (SIPO) –

The shift register, which allows serial input (one bit after the other through a single data line) and produces a parallel output is known as Serial-In Parallel-Out shift register. The circuit consists of four D flip-flops which are connected. The clear (CLR) signal is connected in addition to the clock signal to all the 4 flip flops in order to RESET them. The output of the first flip flop is connected to the input of the next flip flop and so on. All these flip-flops are synchronous with each other since the same clock signal is applied to each flip flop. They are used in communication lines where demultiplexing of a data line into several parallel lines is required because the main use of the SIPO register is to convert serial data into parallel data.



Parallel IN Serial OUT (PISO)

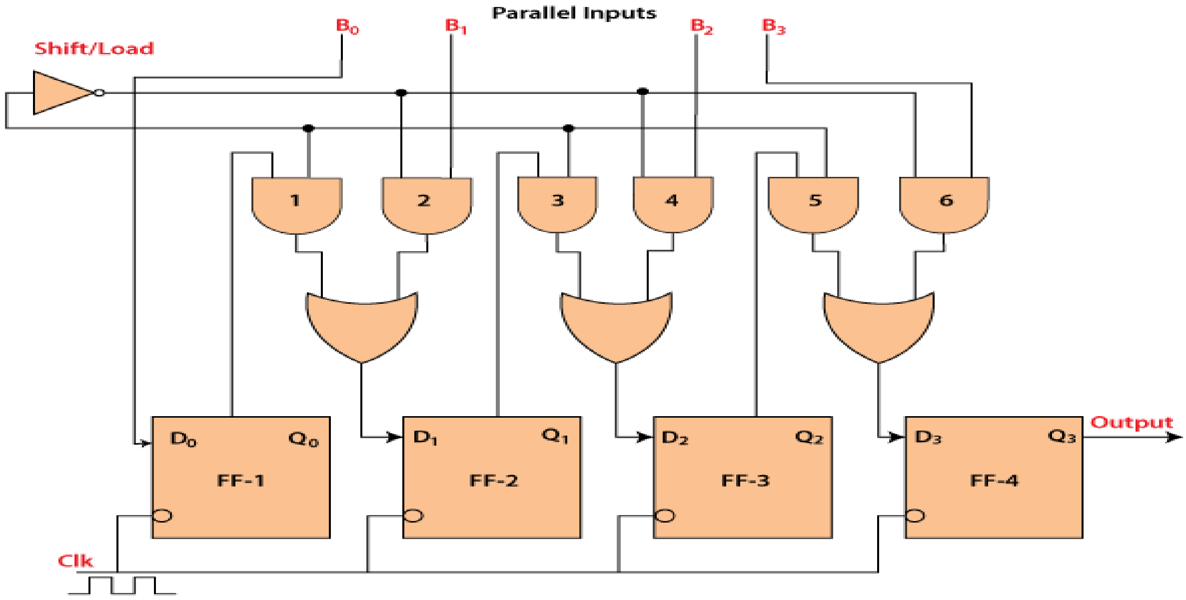
In the **"Parallel IN Serial OUT"** register, the data is entered in a parallel way, and the outcome comes serially. A four-bit **"Parallel IN Serial OUT"** register is designed below. The input of the flip flop is the output of the previous Flip Flop. The input and outputs are connected through the combinational circuit. Through this combinational circuit, the binary input B_0, B_1, B_2, B_3 are passed. The **shift mode** and the **load mode** are the two modes in which the **"PISO"** circuit works.

Load mode

The bits B_0, B_1, B_2, and B_3 are passed to the corresponding flip flops when the second, fourth, and sixth "AND" gates are active. These gates are active when the shift or load bar line set to 0. The binary inputs B0, B1, B2, and B3 will be loaded into the respective flip-flops when the edge of the clock is low. Thus, parallel loading occurs.

Shift mode

The second, fourth, and sixth gates are inactive when the load and shift line set to 0. So, we are not able to load data in a parallel way. At this time, the first, third, and fifth gates will be activated, and the shifting of the data will be left to the right bit. In this way, the **"Parallel IN Serial OUT"** operation occurs.

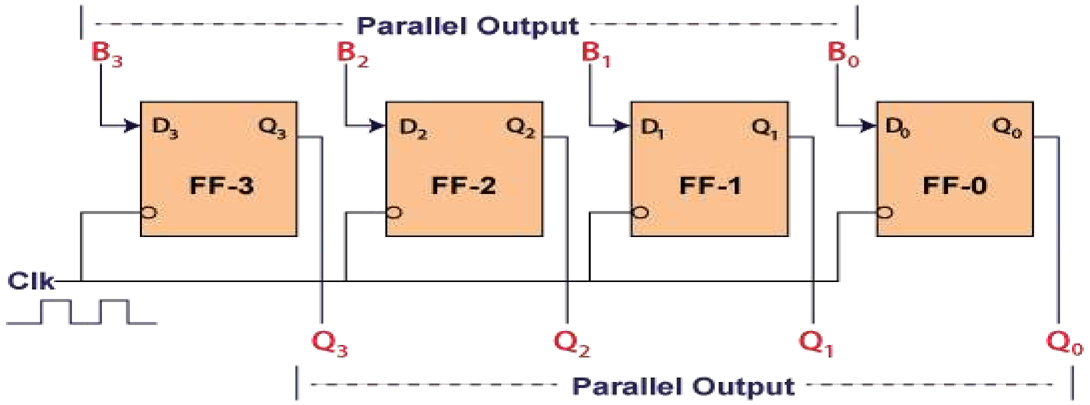


A Parallel in Serial out (PISO) shift register us used to convert parallel data to serial data.

Parallel IN Parallel OUT (PIPO)

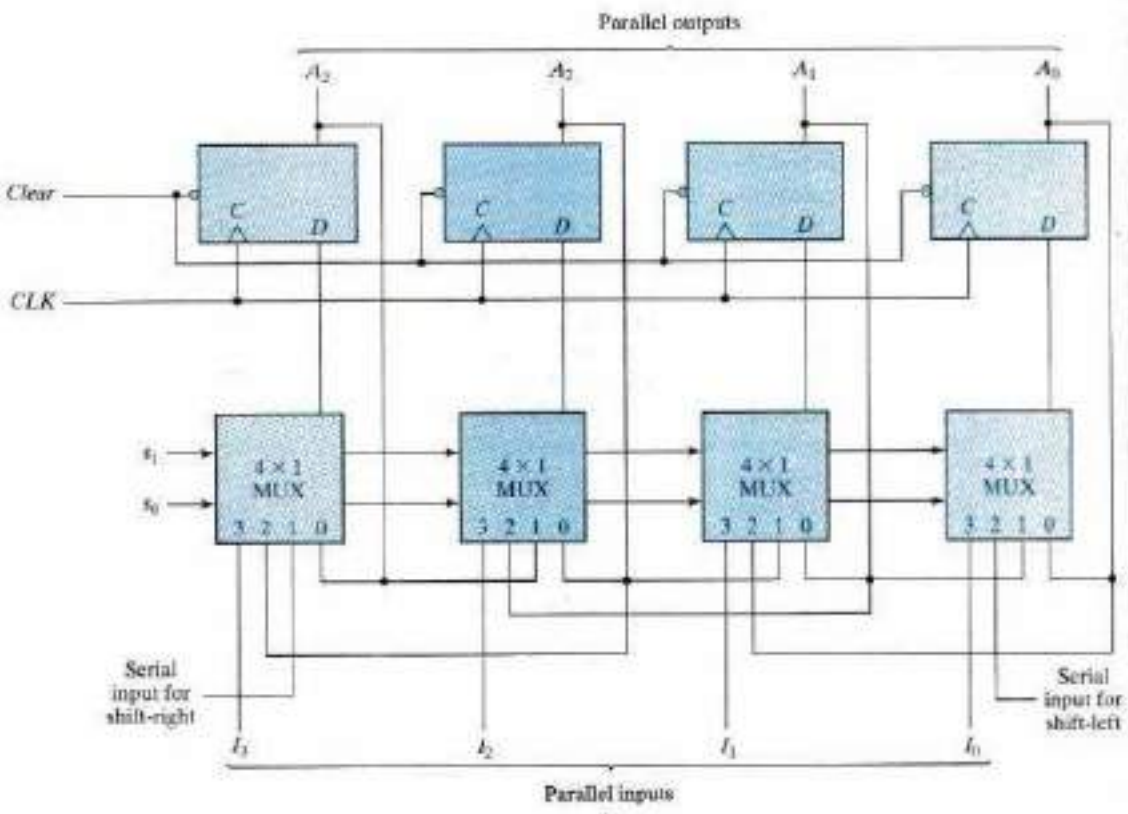
In **"Parallel IN Parallel OUT"**, the inputs and the outputs come in a parallel way in the register.

The inputs A_0, A_1, A_2, and A_3, are directly passed to the data inputs D_0, D_1, D_2, and D_3 of the respective flip flop. The bits of the binary input is loaded to the flip flops when the negative clock edge is applied. The clock pulse is required for loading all the bits. At the output side, the loaded bits appear.



Universal shift register

A Universal shift register is a register which has both the right shift and left shift with parallel load capabilities. Universal shift registers are used as memory elements in computers. A Unidirectional shift register is capable of shifting in only one direction. A bidirectional shift register is capable of shifting in both the directions. The Universal shift register is a combination design of **bidirectional** shift register and a **unidirectional** shift register with parallel load provision



Basic connections –

1. The first input (zeroth pin of multiplexer) is connected to the output pin of the corresponding flip-flop.
2. The second input (first pin of multiplexer) is connected to the output of the veryprevious flip flop which facilitates the right shift.
3. The third input (second pin of multiplexer) is connected to the output of the very-next flip-flop which facilitates the left shift.
4. The fourth input (third pin of multiplexer) is connected to the individual bits of the input data which facilitates parallel loading.

The working of the Universal shift register depends on the inputs given to the select lines.

The register operations performed for the various inputs of select lines are as follows:

S1	S0	Operation
0	0	No change
0	1	Shift right
1	0	Shift left
1	1	Parallel load

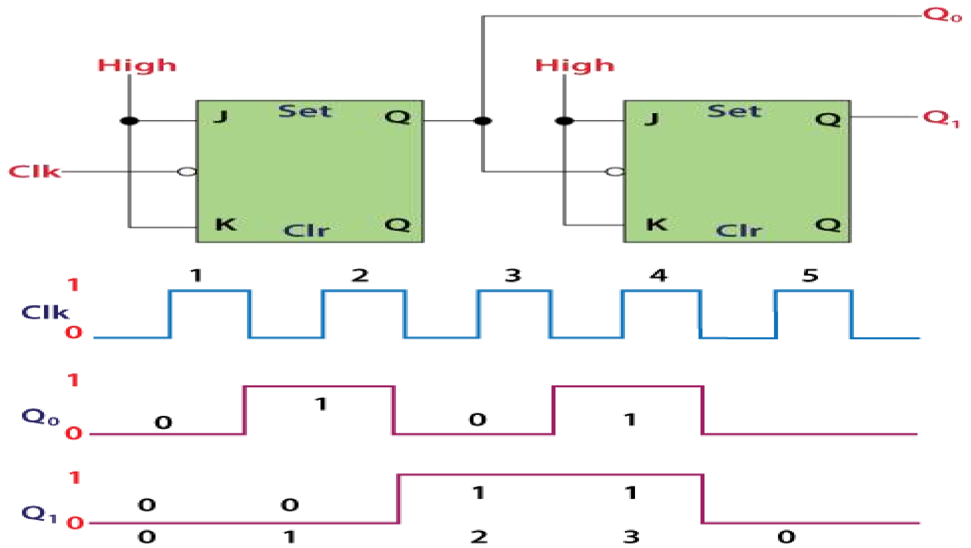
Counters

Counter is a sequential circuit. A digital circuit which is used for a counting pulses is known counter. Counter is the widest application of flip-flops. It is a group of flip-flops with a clock signal applied. Counters are of two types.

- Asynchronous or ripple counters. Synchronous counters.

Asynchronous or ripple counters

The **Asynchronous counter** is also known as the **ripple counter**. Below is a diagram of the 2-bit **Asynchronous counter** in which we used two T flip-flops or two JK flip flopby setting both of the inputs to 1 permanently. The external clock pass to the clock input of the first flip flop, i.e., FF-A and its output, i.e., is passed to clock input of the next flip flop, i.e., FF-B.



Operation:

1. **Condition 1:** When both the flip flops are in reset condition.
   Operation: The outputs of both flip flops, i.e., Q_A Q_B, will be 0.
2. **Condition 2:** When the first negative clock edge passes.
   Operation: The first flip flop will toggle, and the output of this flip flop will change from 0 to 1. The output of this flip flop will be taken by the clock input of the next flip flop. This output will be taken as a positive edge clock by the second flip flop. This input will not change the second flip flop's output state because it is the negative edge triggered flip flop. So, $Q_A = 1$ and $Q_B = 0$
3. **Condition 3:** When the second negative clock edge is applied.
   Operation: The first flip flop will toggle again, and the output of this flip flop will change from 1 to 0. This output will be taken as a negative edge clock by the second flip flop. This input will change the second flip flop's output state because it is the negative edge triggered flip flop. So, $Q_A = 0$ and $Q_B = 1$.
4. **Condition 4:** When the third negative clock edge is applied.
   Operation: The first flip flop will toggle again, and the output of this flip flop will change from 0 to 1. This output will be taken as a positive edge clock by the second flip flop. This input will not change the second flip flop's output state because it is the negative edge triggered flip flop. So, $Q_A = 1$ and $Q_B = 1$
5. **Condition 5:** When the fourth negative clock edge is applied.
   Operation: The first flip flop will toggle again, and the output of this flip flop will change from 1 to 0. This output will be taken as a negative edge clock by the second flip flop. This input will change the output state of the second flip flop. So, $Q_A = 0$ and $Q_B = 0$

Classification of counters

Depending on the way in which the counting progresses, the synchronous or asynchronous counters are classified as follows −

- Up counters
- Down counters
- Up/Down counters

UP/DOWN Counter

Up counter and down counter is combined together to obtain an UP/DOWN counter. A mode control (M) input is also provided to select either up or down mode. A combinational circuit is required to be designed and used between each pair of flip-flop in order to achieve the up/down operation.

- Type of up/down counters
- UP/DOWN ripple counters
- UP/DOWN synchronous counter

UP/DOWN Counter

Up counter and down counter is combined together to obtain an UP/DOWN counter. A mode control (M) input is also provided to select either up or down mode. A combinational circuit is

required to be designed and used between each pair of flip-flop in order to achieve the up/down operation.

- Type of up/down counters
- UP/DOWN ripple counters
- UP/DOWN synchronous counter

UP/DOWN Ripple Counters

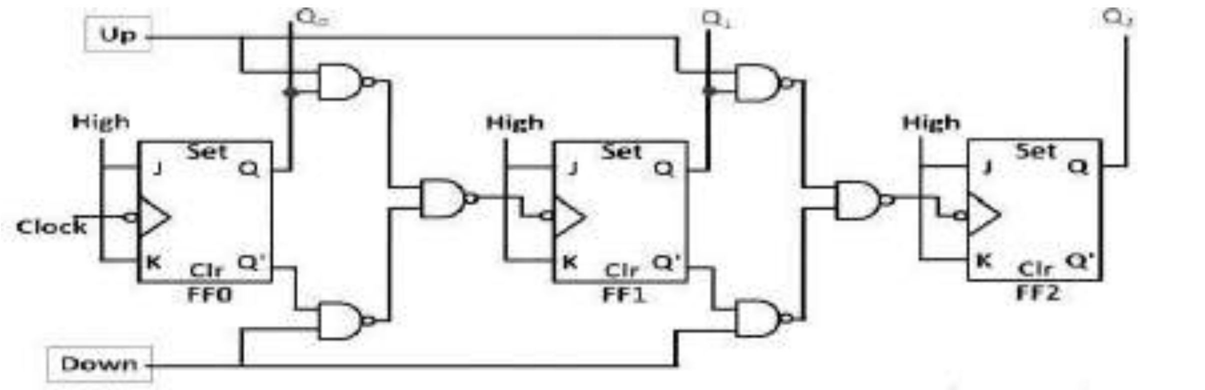
In the UP/DOWN ripple counter all the FFs operate in the toggle mode. So, either T flip-flops or JK flip-flops are to be used. The LSB flip-flop receives clock directly. But the clock to every other FF is obtained from (Q = Q bar) output of the previous FF.

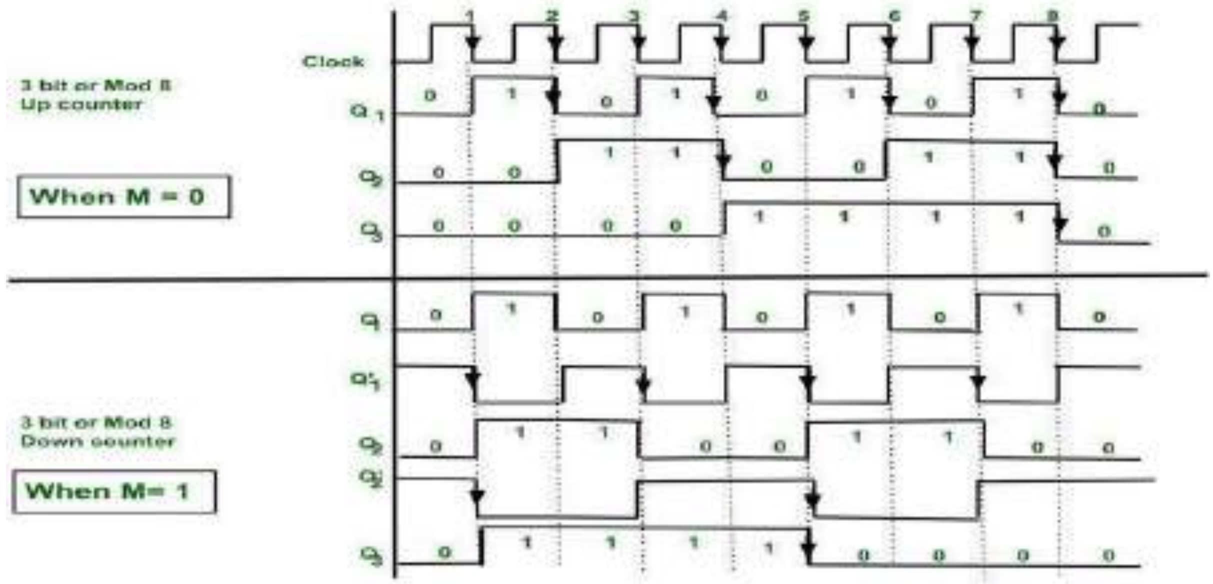
- **UP counting mode (M=0)** − The Q output of the preceding FF is connected to the clock of the next stage if up counting is to be achieved. For this mode, the mode select input M is at logic 0 (M=0).
- **DOWN counting mode (M=1)** − If M = 1, then the Q bar output of the preceding FF is connected to the next FF. This will operate the counter in the counting mode.

Example

3-bit binary up/down ripple counter.

- 3-bit − hence three FFs are required.
- UP/DOWN − So a mode control input is essential.
- For a ripple up counter, the Q output of preceding FF is connected to the clock input of the next one.
- For a ripple up counter, the Q output of preceding FF is connected to the clock input of the next one.
- For a ripple down counter, the Q bar output of preceding FF is connected to the clock input of the next one.
- Let the selection of Q and Q bar output of the preceding FF be controlled by the mode control input M such that, If M = 0, UP counting. So, connect Q to CLK. If M = 1, DOWN counting. So, connect Q bar to CLK.





Modulus Counter (MOD-N Counter)

The 2-bit ripple counter is called as MOD-4 counter and 3-bit ripple counter is called as MOD8 counter. So, in general, an n-bit ripple counter is called as modulo-N counter. Where, MOD number = 2^n.

Type of modulus

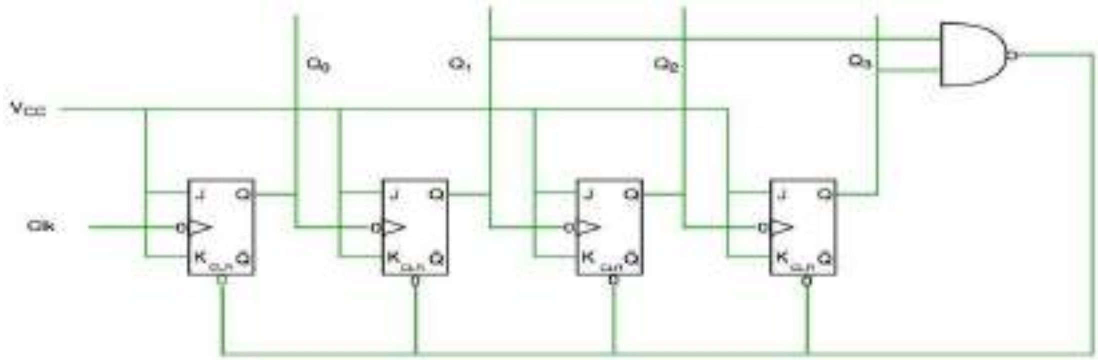
- 2-bit up or down (MOD-4)
- 3-bit up or down (MOD-8)
- 4-bit up or down (MOD-16) **Application of counters**
- Frequency counters
- Digital clock
- Time measurement
- A to D converter
- Frequency divider circuits
- Digital triangular wave generator.

Decade counter

A decade counter counts ten different states and then reset to its initial states. A simple decade counter will count from 0 to 9 but we can also make the decade counters which can go through any ten states between 0 to 15 (for 4 bit counter).

Clock pulse	Q3	Q2	Q1	Q0

0	0	0	0	0
1	0	0	0	1
2	0	0	1	0
3	0	0	1	1
4	0	1	0	0
5	0	1	0	1
6	0	1	1	0
7	0	1	1	1
8	1	0	0	0
9	1	0	0	1
10	0	0	0	0



We see from circuit diagram that we have used NAND gate for Q3 and Q1 and feeding this to clear input line because binary representation of 10 is—1010

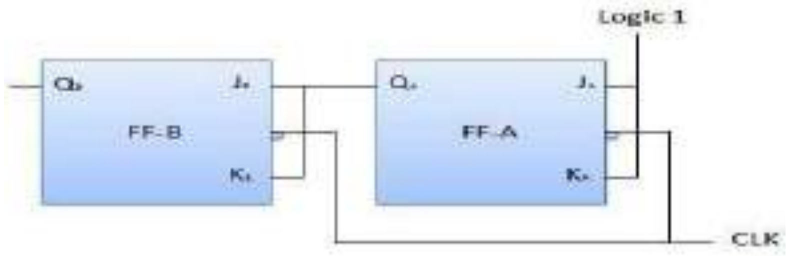
And we see Q3 and Q1 are 1 here, if we give NAND of these two bits to clear input then counter will be clear at 10 and again start from beginning.

Synchronous counters

If the "clock" pulses are applied to all the flip-flops in a counter simultaneously, then such a counter is called as synchronous counter.

2-bit Synchronous up counter

The J_A and K_A inputs of FF-A are tied to logic 1. So, FF-A will work as a toggle flip-flop. The J_B and K_B inputs are connected to Q_A.

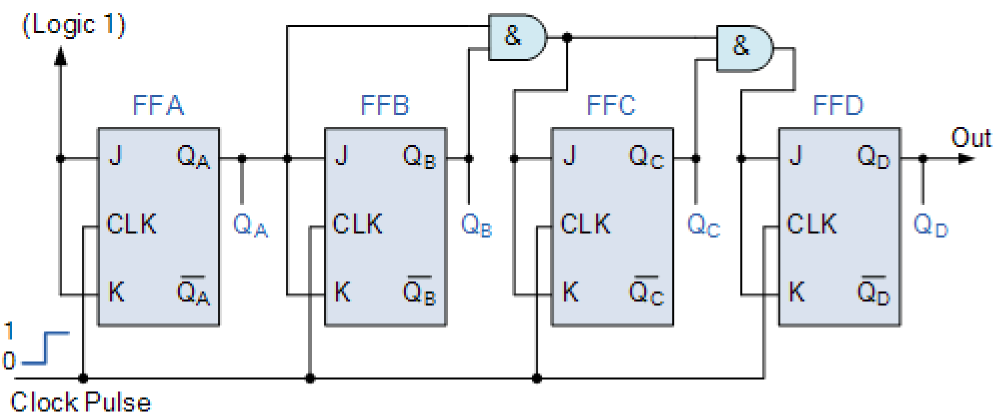


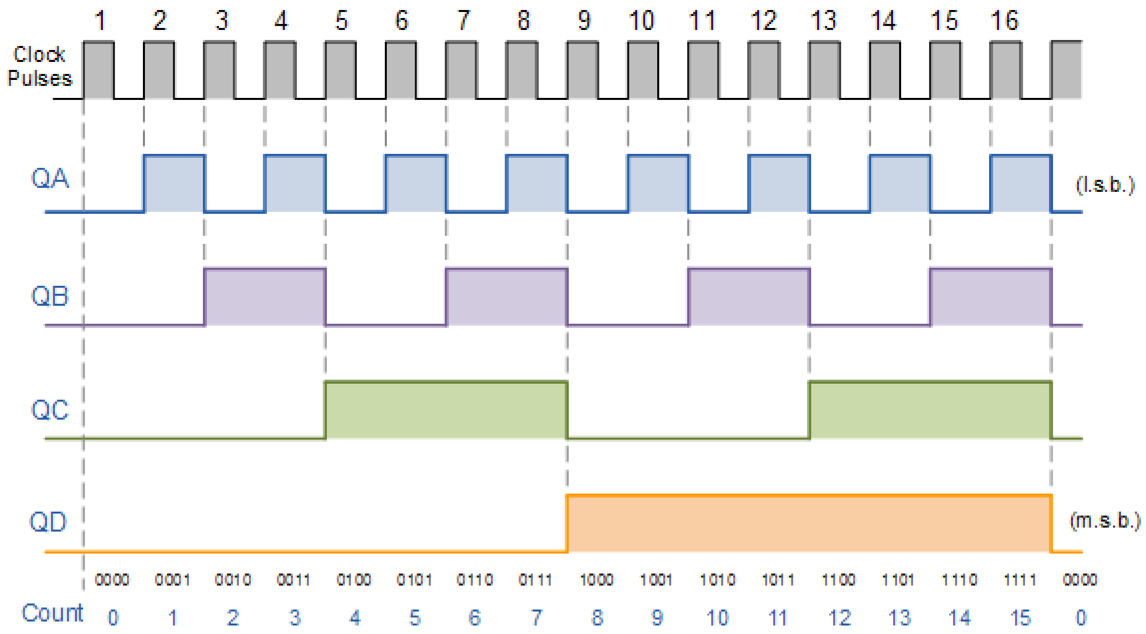
Operation

S.N.	Condition	Operation
1	**Initially let both the FFs be in the reset state**	Q_BQ_A = 00 initially.
2	**After 1st negative clock edge**	As soon as the first negative clock edge is applied, FF-A will toggle and Q_A will change from 0 to 1. But at the instant of application of negative clock edge, Q_A , $J_B = K_B = 0$. Hence FF-B will not change its state. So Q_B will remain 0. Q_BQ_A = 01 after the first clock pulse.
3	**After 2nd negative clock edge**	On the arrival of second negative clock edge, FF-A toggles again and Q_A changes from 1 to 0. But at this instant Q_A was 1. So $J_B = K_B= 1$ and FF-B will toggle. Hence Q_B changes from 0 to 1. Q_BQ_A = 10 after the second clock pulse.
4	**After 3rd negative clock edge**	On application of the third falling clock edge, FF-A will toggle from 0 to 1 but there is no change of state for FF-B. Q_BQ_A = 11 after the third clock pulse.
5	**After 4th negative clock edge**	On application of the next clock pulse, Q_A will change from 1 to 0 as Q_B will also change from 1 to 0. Q_BQ_A = 00 after the fourth clock pulse.

4-bit synchronous counter

- The external clock pulses (pulses to be counted) are fed directly to each of the J-K flipflops in the counter chain and that both the J and K inputs are all tied together in toggle mode, but only in the first flip-flop, flip-flop FFA (LSB) are they connected HIGH, logic
  "1" allowing the flip-flop to toggle on every clock pulse.
- Then the synchronous counter follows a predetermined sequence of states in response to the common clock signal, advancing one state for each pulse.
- The J and K inputs of flip-flop FFB are connected directly to the output Q_A of flipflop FFA, but the J and K inputs of flip-flops FFC and FFD are driven from separate AND gates which are also supplied with signals from the input and output of the previous stage.
- These additional AND gates generate the required logic for the JK inputs of the next stage.
- If we enable each JK flip-flop to toggle based on whether or not all preceding flip-flop outputs (Q) are "HIGH" we can obtain the same counting sequence as with the asynchronous circuit but without the ripple effect, since each flip-flop in this circuit will be clocked at exactly the same time.
- Then as there is no inherent propagation delay in synchronous counters, because all the counter stages are triggered in parallel at the same time, the maximum operating frequency of this type of frequency counter is much higher than that for a similar asynchronous counter circuit.



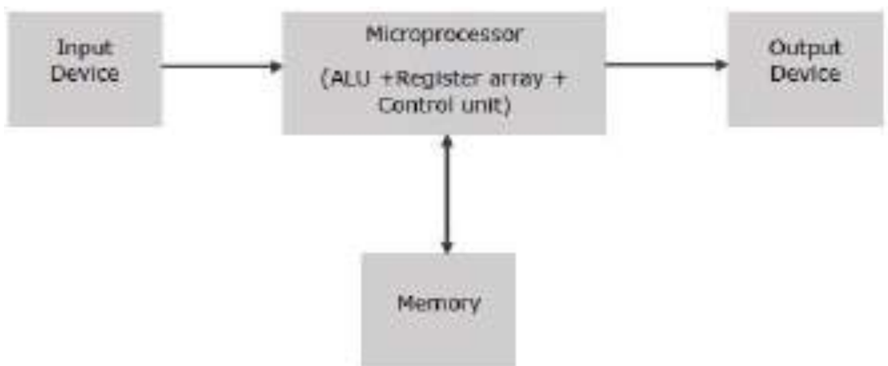


UNIT-4

8085 Microprocessor

- A Microprocessor is a multipurpose, programmable, clock-driven, register-based electronic device that reads binary instructions from a storage device called memory, accepts binary data as input and processes data according to those instructions and provide results as output.
- Microprocessor is a controlling unit of a micro-computer, fabricated on a small chip capable of performing ALU (Arithmetic Logical Unit) operations and communicating with the other devices connected to it.
- Microprocessor consists of an ALU, register array, and a control unit.
- ALU performs arithmetical and logical operations on the data received from the memory or an input device.
- Register array consists of registers identified by letters like B, C, D, E, H, L and accumulator.
- The control unit controls the flow of data and instructions within the computer.

Block Diagram of a Basic Microcomputer



How does a Microprocessor Work?

The microprocessor follows a sequence: Fetch, Decode, and then Execute.

- Initially, the instructions are stored in the memory in a sequential order.
- The microprocessor fetches those instructions from the memory, then decodes it and executes those instructions till STOP instruction is reached.
- Later, it sends the result in binary to the output port. Between these processes, the register stores the temporarily data and ALU performs the computing functions.

Classification of Microprocessor

Microprocessor is classified into two categories-

RISC & CISC

RISC Processor

- RISC stands for Reduced Instruction Set Computer.
- It is designed to reduce the execution time by simplifying the instruction set of the computer.
- Using RISC processors, each instruction requires only one clock cycle to execute results in uniform execution time.
- This reduces the efficiency as there are more lines of code, hence more RAM is needed to store the instructions.
- The compiler also has to work more to convert high-level language instructions into machine code.

Characteristics of RISC

The major characteristics of a RISC processor are as follows −

- It consists of simple instructions.
- It supports various data-type formats.
- It utilizes simple addressing modes and fixed length instructions for pipelining.
- It supports register to use in any context. One cycle execution time.
- “LOAD” and “STORE” instructions are used to access the memory location. It consists of larger number of registers.
- It consists of less number of transistors.

CISC Processor

- CISC stands for Complex Instruction Set Computer.
- It is designed to minimize the number of instructions per program, ignoring the number of cycles per instruction.
- The emphasis is on building complex instructions directly into the hardware.
- The compiler has to do very little work to translate a high-level language into assembly level language/machine code because the length of the code is relatively short, so very little RAM is required to store the instructions.

Characteristics of CISC

- Variety of addressing modes.
- Larger number of instructions.
- Variable length of instruction formats.
- Several cycles may be required to execute one instruction.
- Instruction-decoding logic is complex.
- One instruction is required to support multiple addressing modes.

8085 Microprocessor

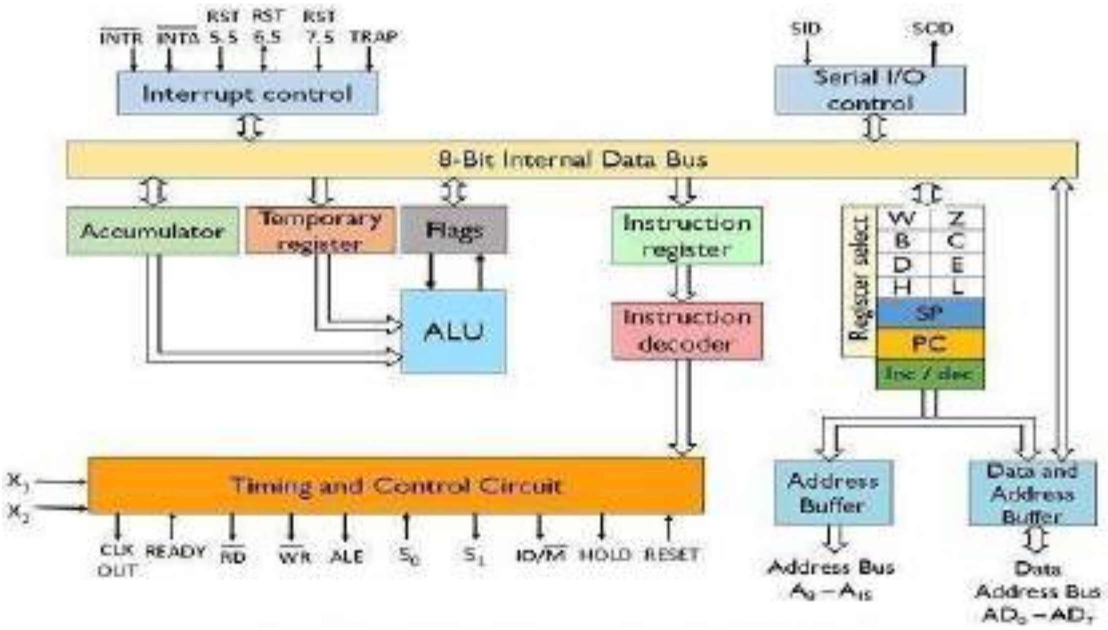
It is an 8-bit microprocessor designed by Intel in 1977 using NMOS technology.

It has the following configuration −

8-bit data bus

- 16-bit address bus, which can address upto 64KB
- A 16-bit program counter
- A 16-bit stack pointer
- Six 8-bit registers arranged in pairs: BC, DE, HL
- Requires +5V supply to operate at 3 MHZ single phase clock It is used in washing machines, microwave ovens, mobile phones, etc.

8085 Microprocessor – Functional Units



8085 consists of the following functional units −

Accumulator

It is an 8-bit register used to perform arithmetic, logical, I/O & LOAD/STORE operations. It is connected to internal data bus & ALU. **Arithmetic and logic unit**

As the name suggests, it performs arithmetic and logical operations like Addition, Subtraction, AND, OR, etc. on 8-bit data.

General purpose register

There are 6 general purpose registers in 8085 processor, i.e., B, C, D, E, H & L. Each register can hold 8-bit data.

These registers can work in pair to hold 16-bit data and their pairing combination is like B-C, D-E & H-L.

Program counter

It is a 16-bit register used to store the memory address location of the next instruction to be executed. Microprocessor increments the program whenever an instruction is being executed, so that the program counter points to the memory address of the next instruction that is going to be executed.

Stack pointer

It is also a 16-bit register works like stack, which is always incremented/decremented by 2 during push & pop operations. **Temporary register**

It is an 8-bit register, which holds the temporary data of arithmetic and logical operations. **Flag register**

It is an 8-bit register having five 1-bit flip-flops, which holds either 0 or 1 depending upon the result stored in the accumulator.

These are the set of 5 flip-flops − Sign
(S)- set to 1 if result is negative.
Zero (Z)- set to 1 if result is zero.
Auxiliary Carry (AC)- set to 1 if carry arises from 3^{rd} bit to 4^{th} bit.
Parity (P)- set to 1 if result has even no. of 1.
Carry (CS)- set to 1 if carry arises after arithmetic and logical operation. Its bit position is shown in the following table –

B7	B6	B5	B4	B3	B2	B1	B0
S	Z	X	AC	X	P	X	CS

Instruction register and decoder

It is an 8-bit register. When an instruction is fetched from memory then it is stored in the Instruction register. Instruction decoder decodes the information present in the Instruction register.

Timing and control unit

It provides timing and control signal to the microprocessor to perform operations. Following are the timing and control signals, which control external and internal circuits − Control Signals: READY, RD’, WR’, ALE
Status Signals: S0, S1, IO/M’
DMA Signals: HOLD, HLDA
RESET Signals: RESET IN, RESET OUT

Interrupt control

As the name suggests it controls the interrupts during a process. When a microprocessor is executing a main program and whenever an interrupt occurs, the microprocessor shifts the control from the main program to process the incoming request. After the request is completed, the control goes back to the main program.

There are 5 interrupt signals in 8085 Microprocessor: INTR, RST 7.5, RST 6.5, RST 5.5, TRAP.

Serial Input/output control

It controls the serial data communication by using these two instructions: SID (Serial input data) and SOD (Serial output data).

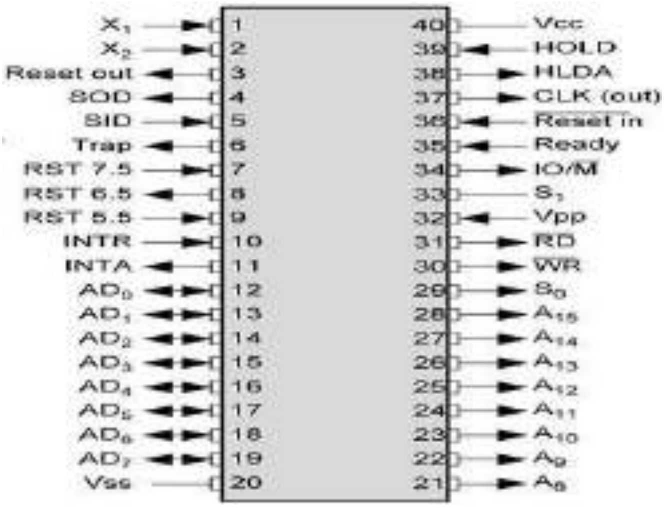
Address buffer and address-data buffer

The content stored in the stack pointer and program counter is loaded into the address buffer and address-data buffer to communicate with the CPU. The memory and I/O chips are connected to these buses; the CPU can exchange the desired data with the memory and I/O chips.

Address bus and data bus

Data bus carries the data to be stored. It is bidirectional, whereas address bus carries the location to where it should be stored and it is unidirectional. It is used to transfer the data & Address I/O devices.

Pin diagram and description



The pins of 8085 microprocessor can be classified into seven groups −

Address bus

A15-A8, it carries the most significant 8-bits of memory/IO address.

Data bus

AD7-AD0, it carries the least significant 8-bit address and data bus.

Control and status signals

These signals are used to identify the nature of operation. There are 3 control signal and 3 status signals.

Three **control signals** are RD, WR & ALE.

RD − This signal indicates that the selected IO or memory device is to be read and is ready for accepting data available on the data bus.

WR − This signal indicates that the data on the data bus is to be written into a selected memory or IO location.

ALE − It is a positive going pulse generated when a new operation is started by the microprocessor. When the pulse goes high, it indicates address. When the pulse goes down it indicates data.

Three **status signals** are IO/M, S0 & S1.

IO/M

This signal is used to differentiate between IO and Memory operations, i.e. when it is high indicates IO operation and when it is low then it indicates memory operation. **S1 & S0**

These signals are used to identify the type of current operation.

Power supply

There are 2 power supply signals − VCC & VSS. **VCC** indicates +5v power supply and **VSS** indicates ground signal.

Clock signals

There are 3 **clock signals**, i.e. X1, X2, CLK OUT.

X1, X2 − A crystal (RC, LC N/W) is connected at these two pins and is used to set frequency of the internal clock generator. This frequency is internally divided by 2.

CLK OUT − This signal is used as the system clock for devices connected with the microprocessor.

Interrupts & externally initiated signals

Interrupts are the signals generated by external devices to request the microprocessor to perform a task. There are 5 interrupt signals, i.e., TRAP, RST 7.5, RST 6.5, RST 5.5, and INTR. We will discuss interrupts in detail in interrupts section.

INTA − It is an interrupt acknowledgment signal.

RESET IN − This signal is used to reset the microprocessor by setting the program counter to zero.

RESET OUT − This signal is used to reset all the connected devices when the microprocessor is reset.

READY − This signal indicates that the device is ready to send or receive data. If READY is low, then the CPU has to wait for READY to go high.

HOLD − This signal indicates that another master is requesting the use of the address and data buses.

HLDA (HOLD Acknowledge) − It indicates that the CPU has received the HOLD request andit will relinquish the bus in the next clock cycle. HLDA is set to low after the HOLD signal is removed.

Serial I/O signals

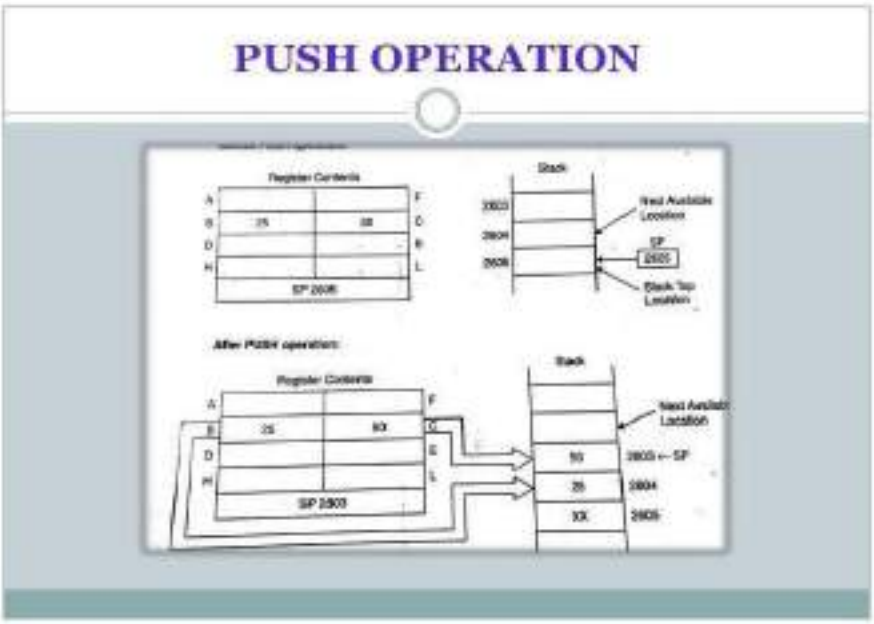
There are 2 serial signals, i.e., SID and SOD and these signals are used for serial communication.

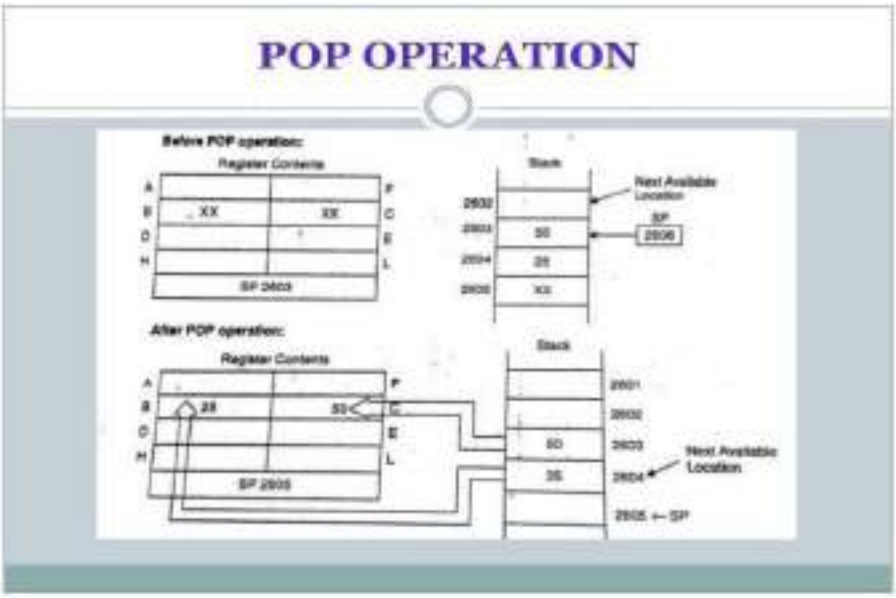
SOD (Serial output data line) − The output SOD is set/reset as specified by the SIM instruction.

SID (Serial input data line) − The data on this line is loaded into accumulator whenever a RIM instruction is executed.

Stack, stack top and stack pointer

- The stack is a LIFO (last in, first out) data structure implemented in the RAM area and is used to store addresses and data when the microprocessor branches to a subroutine.
- Then the return address used to get pushed on this stack.
- Also, to swap values of two registers and register pairs we use the stack as well.
- The Stack Pointer register will hold the address of the top location of the stack.
- On a stack, we can perform two operations.
- PUSH and POP.
- In case of PUSH operation, the SP register gets decreased by 2 and new data item used to insert on to the top of the stack.
- In case of POP operation, the data item will have to be deleted from the top of the stack and the SP register will get increased by the value of 2.





Interrupts

- When microprocessor receives any interrupt signal from peripheral(s) which are requesting its services, it stops its current execution and program control is transferred to a sub-routine by generating CALL signal and after executing sub-routine by generating RET signal again program control is transferred to main program from where it had stopped.
- When microprocessor receives interrupt signals, it sends an acknowledgement (INTA) to the peripheral which is requesting for its service.

Interrupts can be classified into various categories based on different parameters:

1. Hardware and Software Interrupts –

When microprocessors receive interrupt signals through pins (hardware) of microprocessor, they are known as **Hardware Interrupts**. There are 5 Hardware Interrupts in 8085 Microprocessor. They are – INTR, RST 7.5, RST 6.5, RST 5.5, TRAP.

Software Interrupts are those which are inserted in between the program which means these are mnemonics of microprocessor. There are 8 software interrupts in 8085 Microprocessor.

They are – RST 0, RST 1, RST 2, RST 3, RST 4, RST 5, RST 6, RST 7.

2. Vectored and Non-Vectored Interrupts –

Vectored Interrupts are those which have fixed vector address (starting address of subroutine) and after executing these, program control is transferred to that address.

Vector Addresses are calculated by the formula 8 * TYPE

INTERRUPT	VECTOR ADDRESS
TRAP (RST 4.5)	24 H
RST 5.5	2C H
RST 6.5	34 H
RST 7.5	3C H

For Software interrupts vector addresses are given by:

INTERRUPT	VECTOR ADDRESS
RST 0	00 H
RST 1	08 H
RST 2	10 H
RST 3	18 H
RST 4	20 H
RST 5	28 H
RST 6	30 H
RST 7	38 H

Non-Vectored Interrupts are those in which vector address is not predefined. The interrupting device gives the address of sub-routine for these interrupts. INTR is the only nonvectored interrupt in 8085 Microprocessor.

3. Maskable and Non-Maskable Interrupts –

Maskable Interrupts are those which can be disabled or ignored by the microprocessor. These interrupts are either edge-triggered or level-triggered, so they can be disabled. INTR, RST 7.5, RST 6.5, RST 5.5 are maskable interrupts in 8085 Microprocessor.

Non-Maskable Interrupts are those which cannot be disabled or ignored by microprocessor. TRAP is a non-maskable interrupt. It consists of both level as well as edge triggering and is used in critical power failure conditions.

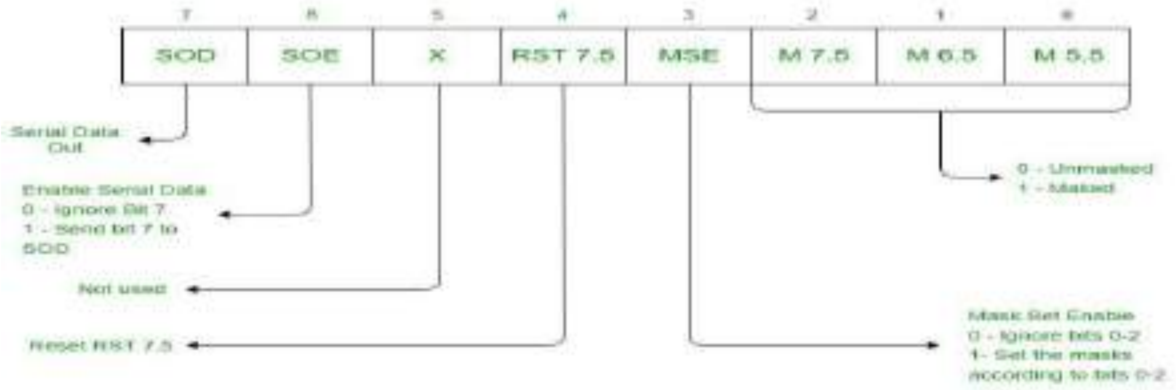
Priority of Interrupts –

When microprocessor receives multiple interrupt requests simultaneously, it will execute the interrupt service request (ISR) according to the priority of the interrupts.

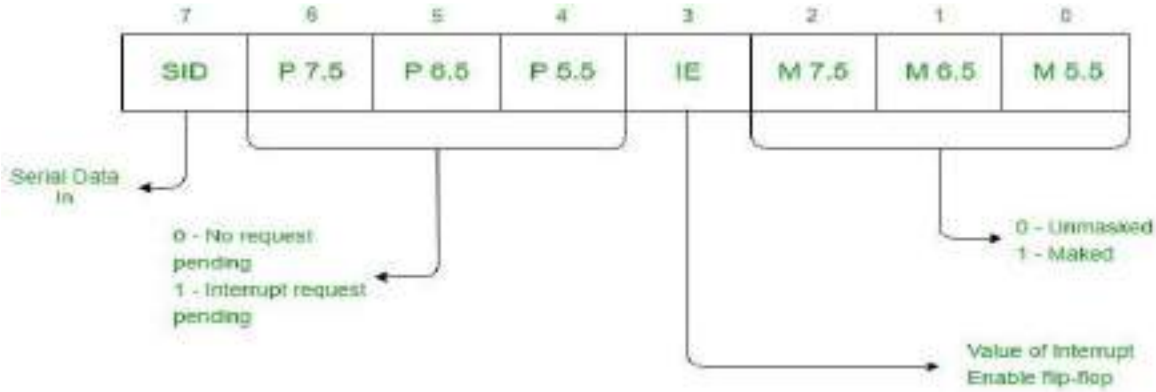
Instruction for Interrupts –

i. Enable Interrupt (EI) – The interrupt enable flip-flop is set and all interrupts are enabled following the execution of next instruction followed by EI. No flags are affected. After a system reset, the interrupt enable flip-flop is reset, thus disabling the interrupts. This instruction is necessary to enable the interrupts again (except TRAP). **ii. Disable Interrupt (DI) –** This instruction is used to reset the value of enable flip-flop hence disabling all the interrupts. No flags are affected by this instruction.

iii. Set Interrupt Mask (SIM) – It is used to implement the hardware interrupts (RST 7.5, RST 6.5, RST 5.5) by setting various bits to form masks or generate output data via the Serial Output Data (SOD) line. First the required value is loaded in accumulator then SIM will take the bit pattern from it.



iv. Read Interrupt Mask (RIM) – This instruction is used to read the status of the hardware interrupts (RST 7.5, RST 6.5, RST 5.5) by loading into the A register a byte which defines the condition of the mask bits for the interrupts. It also reads the condition of SID (Serial Input Data) bit on the microprocessor.



Opcodes and operands

- Instruction is divided into two parts: opcodes and operands.
- The opcode is the instruction that is executed by the CPU and the operand is the data or memory location used to execute that instruction.
- An operand (written using hexadecimal notation) provides the data itself, or the location where the data to be processed is stored.
- Some instructions do not require an operand and some may require more than one operand.

Instruction size

- The 8085 instruction set is classified into 3 categories by considering the length of the instructions.
- Three types of instruction are: 1-byte instruction, 2-byte instruction, and 3-byte instruction.

1. One-byte instructions –
In 1-byte instruction, the opcode and the operand of an instruction are represented in one byte.
Example- MOV A,B

2. Two-byte instructions –
Two-byte instruction is the type of instruction in which the first 8 bits indicates the opcode and the next 8 bits indicates the operand.
Example- MVI A,34H
3. Three-byte instructions –
Three-byte instruction is the type of instruction in which the first 8 bits indicates the opcode and the next two bytes specify the 16-bit address. The low-order address is represented in second byte and the high-order address is represented in the third byte. Example- LDA 2000H

Instruction set of 8085 Microprocessor

8085 instruction set is classified in 5 groups- Data transfer group
Arithmetic group
Logical group
Branch control group
Machine control group

Data transfer group

Data transfer instructions are the instructions which transfers data in the microprocessor. They are also called copy instructions.

Opcode	Operand	Explanation	Example
MOV	R1,R2	Move the data from R2 to R1	MOV A,B
MOV	R,M	Move data from memory location to R	MOV B,M
MVI	R,8-bit data	Move the immediate 8-bit data to R	MVI C,34H
MVI	M,8-bit data	Move the immediate 8-bit data to memory location	MVI M,23H
LDA	16-bit address	Load the data from 16-bit address to ACC	LDA 2000H
STA	16-bit address	Store the data of ACC to 16-bit address	STA 2500H
LHLD	16-bit address	Directly loads at H & L registers	LHLD 2050
SHLD	16-bit address	directly stores from H & L registers	SHLD 2050
LXI	rp, 16-bit data	loads the specified register pair with data	LXI H, 3050
XCHG		exchanges H with D, and L with E	XCHG
PUSH	rp	pushes rp to the stack	PUSH H
POP	rp	pops the stack to rp	POP H
IN	8-bit port address	inputs contents of the specified port to A	IN 01
OUT	8-bit port address	outputs contents of A to the specified port	OUT 02

Arithmetic group

Arithmetic Instructions are the instructions which perform basic arithmetic operations such as addition, subtraction and a few more. In 8085 Microprocessor, the destination operand is generally the accumulator. In 8085 Microprocessor, the destination operand is generally the accumulator.

Opcode	Operand	Explanation	Example
ADD	R	A = A + R	ADD B
ADD	M	A = A + M	ADD M
ADI	8-bit data	A = A + 8-bit data	ADI 50
ADC	R	A = A + R + prev. carry	ADC B
ADC	M	A = A + Mc + prev. carry	ADC M
ACI	8-bit data	A = A + 8-bit data + prev. carry	ACI 50
SUB	R	A = A – R	SUB B
SUB	M	A = A – M	SUB M
SUI	8-bit data	A = A – 8-bit data	SUI 50
SBB	R	A = A – R – prev. carry	SBB B
SBB	M	A = A – M -prev. carry	SBB M
SBI	8-bit data	A = A – 8-bit data – prev. carry	SBI 50
INR	R	R = R + 1	INR B
INR	M	M = M + 1	INR M
INX	r.p.	r.p. = r.p. + 1	INX H

DCR	R	R = R – 1	DCR B
DCR	M	M = M – 1	DCR M
DCX	r.p.	r.p. = r.p. – 1	DCX H
DAD	r.p.	HL = HL + r.p.	DAD H

Logical group

Logical instructions are the instructions which perform basic logical operations such as AND, OR, etc. In 8085 Microprocessor, the destination operand is always the accumulator. Here logical operation works on a bitwise level.

Opcode	Operand	Explanation	Example
ANA	R	A = A AND R	ANA B
ANA	M	A = A AND M	ANA M
ANI	8-bit data	A = A AND 8-bit data	ANI 50
ORA	R	A = A OR R	ORA B
ORA	M	A = A OR M	ORA M
ORI	8-bit data	A = A OR 8-bit data	ORI 50
XRA	R	A = A XOR R	XRA B
XRA	M	A = A XOR M	XRA M
XRI	8-bit data	A = A XOR 8-bit data	XRI 50
CMA		A = 1's compliment of A	CMA
CMP	R	Compares R with A and triggers the flag register	CMP B
CMP	M	Compares Mc with A and triggers the flag register	CMP M
CPI	8-bit data	Compares 8-bit data with A and triggers the flag register	CPI 50
RRC		Rotate accumulator right without carry	RRC
RLC		Rotate accumulator left without carry	RLC
RAR		Rotate accumulator right with carry	RAR
RAL		Rotate accumulator left with carry	RAR
CMC		Compliments the carry flag	CMC
STC		Sets the carry flag	STC

Branch group

Branching instructions refer to the act of switching execution to a different instruction sequence as a result of executing a branch instruction. The three types of branching instructions are:

1. Jump (unconditional and conditional)
2. Call (unconditional and conditional)
3. Return (unconditional and conditional)

1. Jump Instructions – The jump instruction transfers the program sequence to the memory address given in the operand based on the specified flag. Jump instructions are 2 types: Unconditional Jump Instructions and Conditional Jump Instructions.

(a) Unconditional Jump Instructions: Transfers the program sequence to the described memory address.

JMP 16-bit address   Jumps to the address

Example- JMP 2050

(b) Conditional Jump Instructions: Transfers the program sequence to the described memory address only if the condition in satisfied.

Opcode	Operand	Explanation	Example
JC	Address	Jumps to the address if carry flag is 1	JC 2050
JNC	Address	Jumps to the address if carry flag is 0	JNC 2050

JZ	Address	Jumps to the address if zero flag is 1	JZ 2050
JNZ	Address	Jumps to the address if zero flag is 0	JNZ 2050
JPE	Address	Jumps to the address if parity flag is 1	JPE 2050
JPO	Address	Jumps to the address if parity flag is 0	JPO 2050
JM	Address	Jumps to the address if sign flag is 1	JM 2050
JP	Address	Jumps to the address if sign flag 0	JP 2050
JC	Address	Jumps to the address if carry flag is 1	JC 2050
JNC	Address	Jumps to the address if carry flag is 0	JNC 2050
JZ	Address	Jumps to the address if zero flag is 1	JZ 2050
JNZ	Address	Jumps to the address if zero flag is 0	JNZ 2050
JPE	Address	Jumps to the address if parity flag is 1	JPE 2050
JPO	Address	Jumps to the address if parity flag is 0	JPO 2050
JM	Address	Jumps to the address if sign flag is 1	JM 2050
JP	Address	Jumps to the address if sign flag 0	JP 2050

2. Call Instructions – The call instruction transfers the program sequence to the memory address given in the operand. Before transferring, the address of the next instruction after CALL is pushed onto the stack. Call instructions are 2 types: Unconditional Call Instructions and Conditional Call Instructions.

(a) Unconditional Call Instructions: It transfers the program sequence to the memory address given in the operand.

CALL 16-address Unconditionally calls

Example- CALL 2050

(b) Conditional Call Instructions: Only if the condition is satisfied, the instructions executes.

Opcode	Operand	Explanation	Example
CC	Address	Call if carry flag is 1	CC 2050
CNC	Address	Call if carry flag is 0	CNC 2050
CZ	Address	Calls if zero flag is 1	CZ 2050
CNZ	Address	Calls if zero flag is 0	CNZ 2050
CPE	Address	Calls if parity flag is 1	CPE 2050
CPO	Address	Calls if parity flag is 0	CPO 2050
CM	Address	Calls if sign flag is 1	CM 2050
CP	Address	Calls if sign flag is 0	CP 2050

3. Return Instructions – The return instruction transfers the program sequence from the subroutine to the calling program. Return instructions are 2 types: Unconditional Jump Instructions and Conditional Jump Instructions.

(a) Unconditional Return Instruction: The program sequence is transferred unconditionally from the subroutine to the calling program. RET Return from the subroutine unconditionally

(b) Conditional Return Instruction: The program sequence is transferred unconditionally from the subroutine to the calling program only is the condition is satisfied.

Opcode	Operand	Explanation	Example
RC		Return from the subroutine if carry flag is 1	RC

RNC		Return from the subroutine if carry flag is 0	RNC
RZ		Return from the subroutine if zero flag is 1	RZ
RNZ		Return from the subroutine if zero flag is 0	RNZ
RPE		Return from the subroutine if parity flag is 1	RPE
RPO		Return from the subroutine if parity flag is 0	RPO
RM		Returns from the subroutine if sign flag is 1	RM
RP		Returns from the subroutine if sign flag is 0	RP

Machine control group

Opcode	**Operand**	**Meaning**	**Explanation**
NOP		No operation	No operation is performed, i.e., the instruction is fetched and decoded.
HLT		Halt and enter wait state	The CPU finishes executing the current instruction and stops further execution. An interrupt or reset is necessary to exit from the halt state.
DI		Disable interrupts	The interrupt enable flip-flop is reset and all the interrupts are disabled except TRAP.
EI		Enable interrupts	The interrupt enable flip-flop is set and all the interrupts are enabled.
RIM		Read interrupt mask	This instruction is used to read the status of interrupts 7.5, 6.5, 5.5 and read serial data input bit.
SIM		Set interrupt mask	This instruction is used to implement the interrupts 7.5, 6.5, 5.5, and serial data output.

Addressing modes

The term addressing modes refers to the way in which the operand of an instruction is specified.

Types of addressing modes –

In 8085 microprocessor there are 5 types of addressing modes:

Immediate Addressing Mode –

In immediate addressing mode the source operand is always data. If the data is 8-bit, then the instruction will be of 2 bytes, if the data is of 16-bit then the instruction will be of 3 bytes.

Examples:

MVI B,45 (move the data 45H immediately to register B)

LXI H,3050 (load the H-L pair with the operand 3050H immediately)

JMP address (jump to the operand address immediately)

Register Addressing Mode –

In register addressing mode, the data to be operated is available inside the register(s) and register(s) is(are) operands. Therefore, the operation is performed within various registers of the microprocessor.

Examples:

MOV A, B (move the contents of register B to register A)

ADD B (add contents of registers A and B and store the result in register A) INR A (increment the contents of register A by one)

Direct Addressing Mode –

In direct addressing mode, the data to be operated is available inside a memory location and that memory location is directly specified as an operand. The operand is directly available in the instruction itself.

Examples:

LDA 2050 (load the contents of memory location into accumulator A)

LHLD address (load contents of 16-bit memory location into H-L register pair)

IN 35 (read the data from port whose address is 35)

Register Indirect Addressing Mode –

In register indirect addressing mode, the data to be operated is available inside a memory location and that memory location is indirectly specified by a register pair.

Examples:

MOV A, M (move the contents of the memory location pointed by the H-L pair to the accumulator)

LDAX B (move contents of B-C register to the accumulator)

LHLD 9570 (load immediate the H-L pair with the data of the location 9570)

Implied/Implicit Addressing Mode –

In implied/implicit addressing mode the operand is hidden and the data to be operated is available in the instruction itself. Examples:

CMA (finds and stores the 1’s complement of the contents of accumulator A in A)

RRC (rotate accumulator A right by one bit)

RLC (rotate accumulator A left by one bit)

Instruction cycle of 8085 Microprocessor

Time required to execute and fetch an entire instruction is called instruction cycle. It consists:

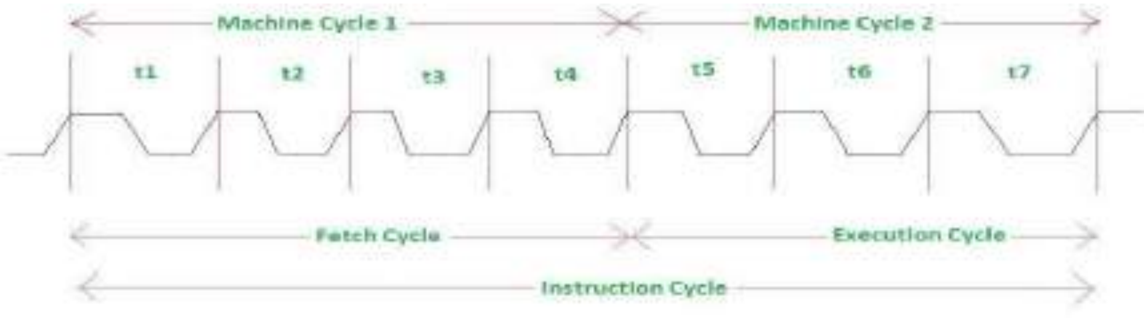
Fetch cycle – The next instruction is fetched by the address stored in program counter (PC) and then stored in the instruction register.

Decode instruction – Decoder interprets the encoded instruction from instruction register.

Execution cycle – consists memory read (MR), memory write (MW), input output read (IOR) and input output write (IOW)

The time required by the microprocessor to complete an operation of accessing memory or input/output devices is called **machine cycle**. One time period of frequency of microprocessor is called **t-state**. A t-state is measured from the falling edge of one clock pulse to the falling edge of the next clock pulse.

Fetch cycle takes four t-states and execution cycle takes three t-states.

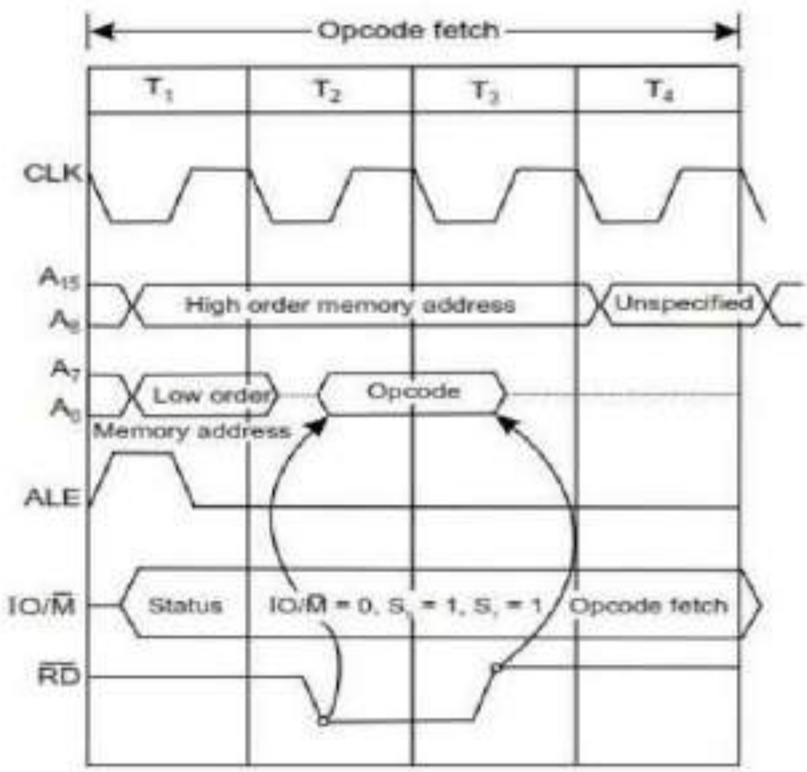


Instruction cycle in 8085 microprocessor

Timing diagram

Timing Diagram is a graphical representation. It represents the execution time taken by each instruction in a graphical format. The execution time is represented in T-states.

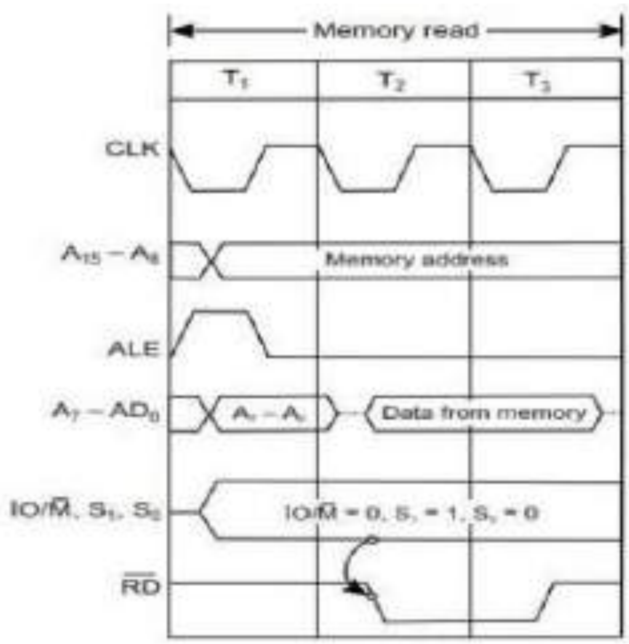
Opcode fetch cycle



- Each instruction of the processor has one byte opcode.
- The opcodes are stored in memory. So, the processor executes the opcode fetch machine cycle to fetch the opcode from memory.
- Hence, every instruction starts with opcode fetch machine cycle.
- The time taken by the processor to execute the opcode fetch cycle is 4T.
- In this time, the first, 3 T-states are used for fetching the opcode from memory and the remaining T-states are used for internal operations by the processor.

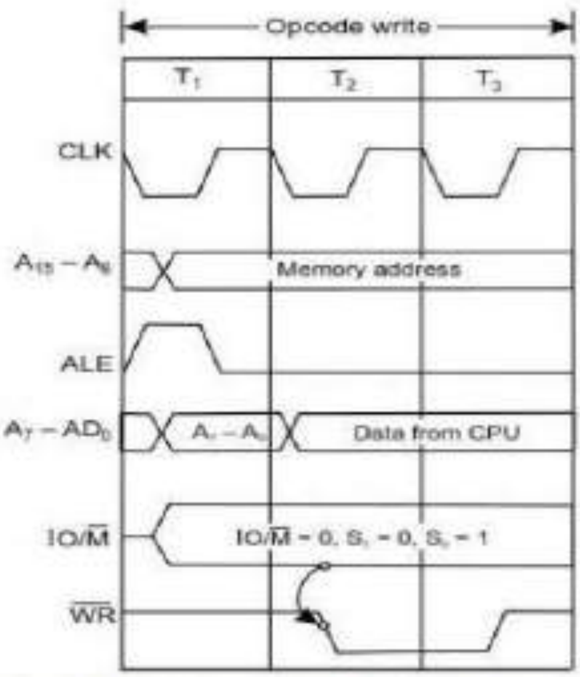
Memory read cycle

- The memory read machine cycle is executed by the processor to read a data byte from memory.
- The processor takes 3T states to execute this cycle.
- The instructions which have more than one byte word size will use the machine cycle after the opcode fetch machine cycle.



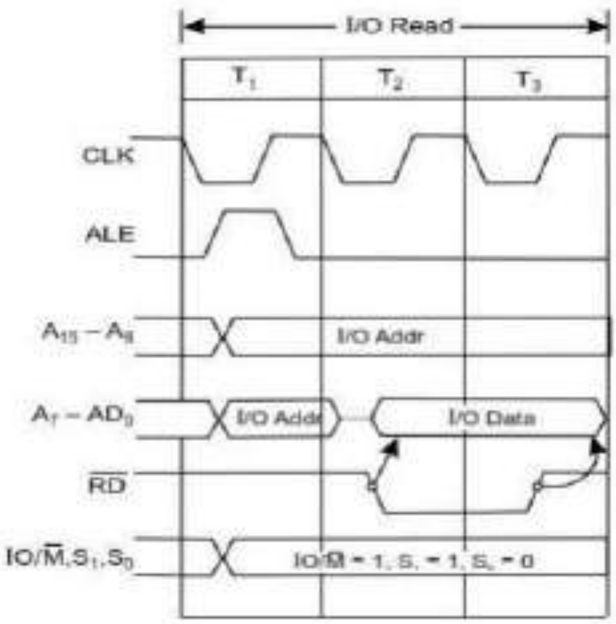
Memory write cycle

- The memory write machine cycle is executed by the processor to write a data byte in a memory location.
- The processor takes, 3T states to execute this machine cycle.



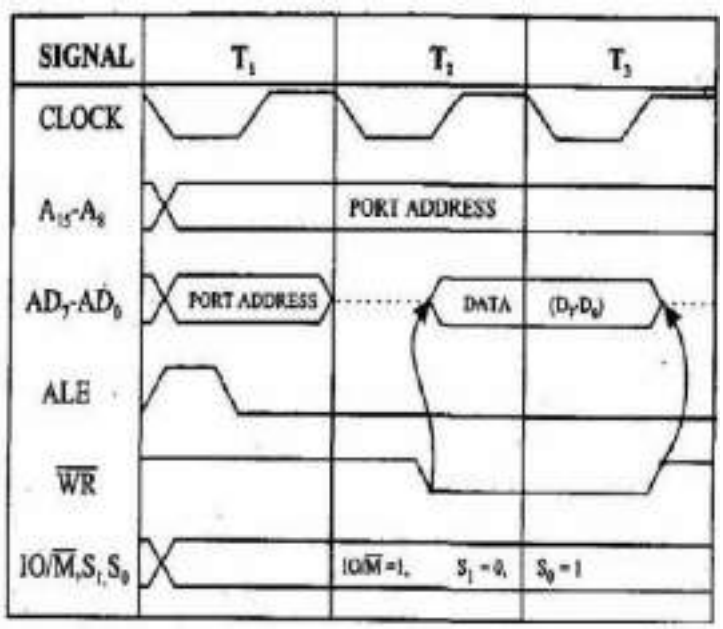
I/O read cycle

- The I/O Read cycle is executed by the processor to read a data byte from I/O port or from the peripheral, which is I/O, mapped in the system.
- The processor takes 3T states to execute this machine cycle.
- The IN instruction uses this machine cycle during the execution.



I/O write cycle

- The I/O Read cycle is executed by the processor to write a data byte from system to I/O port or peripheral, which is I/O mapped.
- The processor takes 3T states to execute this machine cycle.
- The OUT instruction uses this machine cycle during the execution.

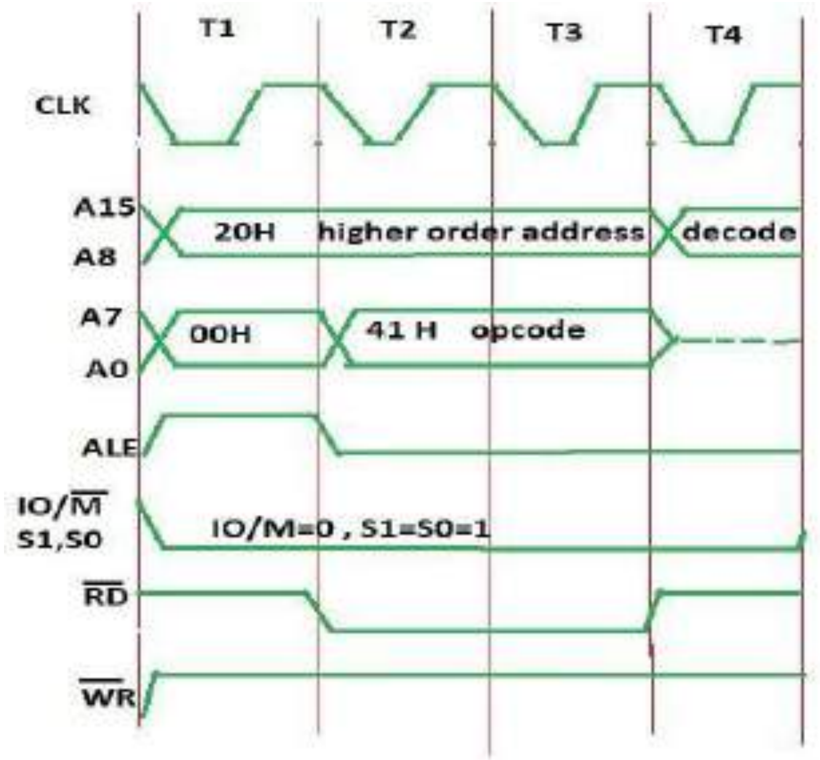


Example-1

The instruction MOV B, C is of 1 byte; therefore, the complete instruction will be stored in a single memory address.

2000 MOV B,C

Only opcode fetching is required for this instruction and thus we need 4 T states for the timing diagram. For the opcode fetch the IO/M (low active) = 0, S1 = 1 and S0 = 1.



In Opcode fetch (t1-t4 T-states):

1. 00 – lower bit of address where opcode is stored, i.e., 00 2.
20 – higher bit of address where opcode is stored, i.e., 20.
3. ALE – provides signal for multiplexed address and data bus. Only in t1 it used as address bus to fetch lower bit of address otherwise it will be used as data bus.
4. RD (low active) – signal is 1 in t1 & t4 as no data is read by microprocessor. Signal is 0 in t2 & t3 because here the data is read by microprocessor.
5. WR (low active) – signal is 1 throughout, no data is written by microprocessor.
6. IO/M (low active) – signal is 1 in throughout because the operation is performing on memory.
7. S0 and S1 – both are 1 in case of opcode fetching.

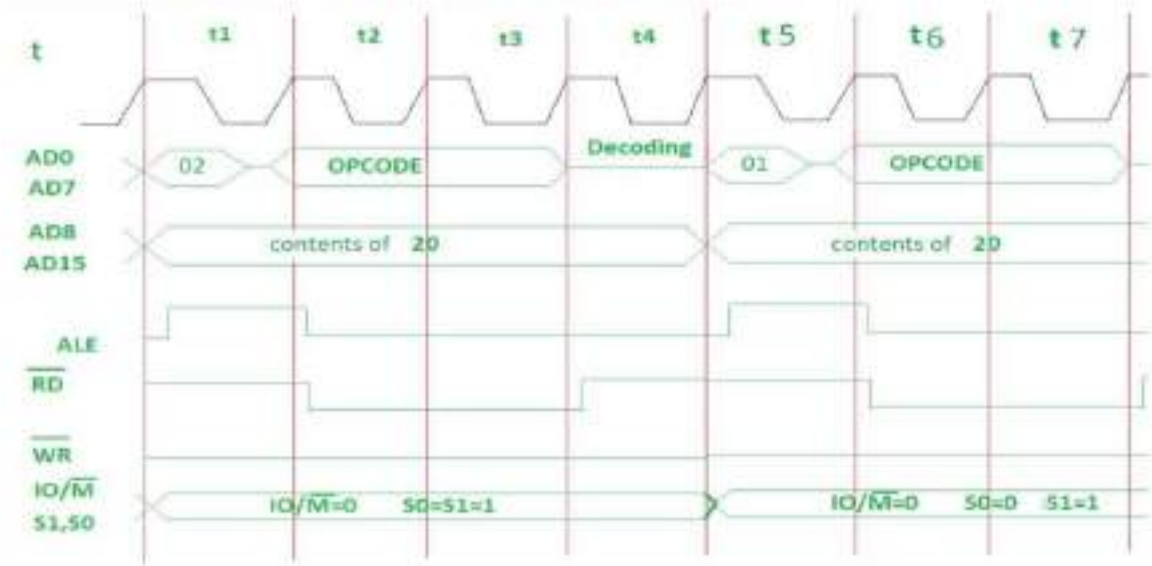
Example-2

MVI B, 45

2000: Opcode

2001: 45

- The opcode fetch will be same in all the instructions.
- Only the read instruction of the opcode needs to be added in the successive T states.
- For the opcode fetch the IO/M (low active) = 0, S1 = 1 and S0 = 1. Also, 4 T states will be required to fetch the opcode from memory.
- For the opcode read the IO/M (low active) = 0, S1 = 1 and S0 = 0. Also, only 3 T states will be required to read data from memory.



In Opcode fetch (t1-t4 T-states) –

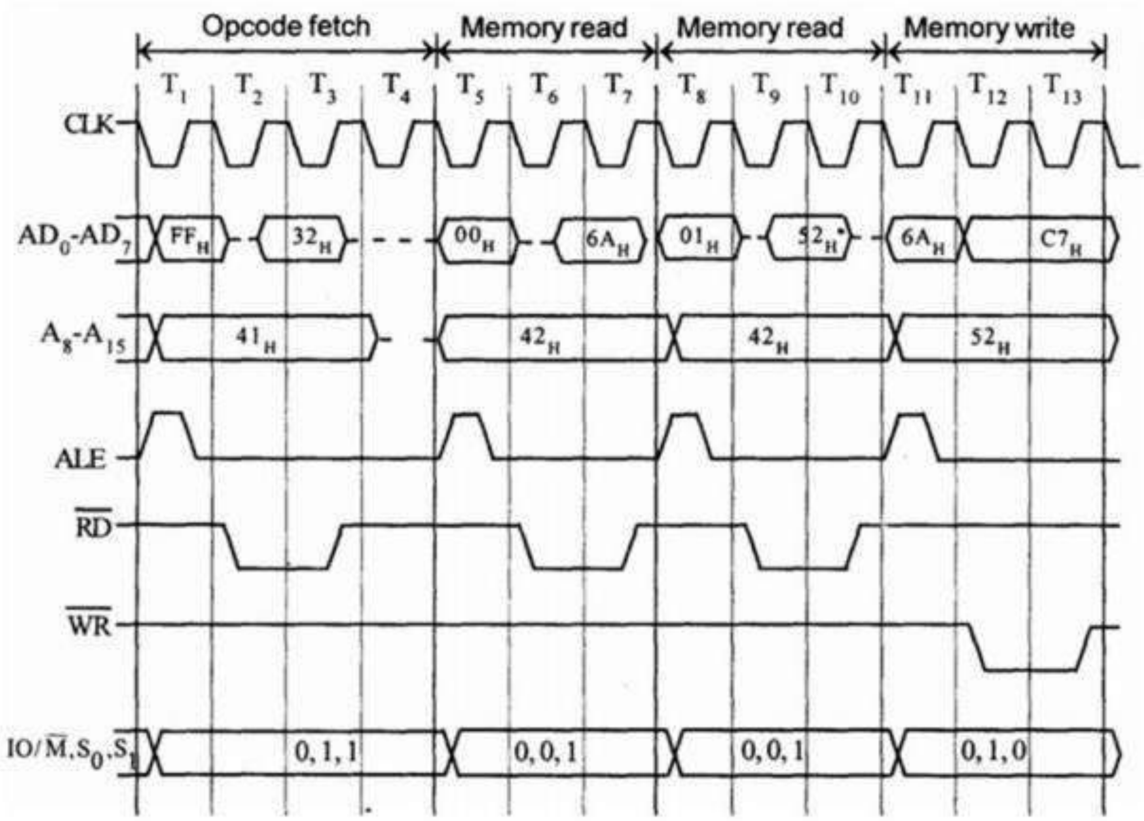
1. 00 – lower bit of address where opcode is stored.
2. 20 – higher bit of address where opcode is stored.
3. ALE – Provides signal for multiplexed address and data bus. Only in t1 it used as address bus to fetch lower bit of address otherwise it will be used as data bus.
4. RD (low active) – Signal is 1 in t1 & t4, no data is read by microprocessor. Signal is 0 in t2 & t3, data is read by microprocessor.
5. WR (low active) – Signal is 1 throughout, no data is written by microprocessor.
6. IO/M (low active) – Signal is 0 in throughout, operation is performing on memory.
7. S0 and S1 – Signal is 1 in t1 to t4 states, as to fetch the opcode from the memory.

In Opcode read (t5-t7 T-states) –

1. 01 – lower bit of address where data is stored.
2. 320 – higher bit of address where data is stored.
3. ALE – Provides signal for multiplexed address and data bus. Only in t5 it used as address bus to fetch lower bit of address otherwise it will be used as data bus.
4. RD (low active) – Signal is 1 in t5 as no data is read by microprocessor. Signal is 0 in t6 & t7 as data is read by microprocessor.
5. WR (low active) – Signal is 1 throughout, no data is written by microprocessor.
6. IO/M (low active) – Signal is 0 in throughout, operation is performing on memory.
7. S0 – Signal is 0 in throughout, operation is performing on memory to read data 45.
8. S1 – Signal is 1 throughout, operation is performing on memory to read data 45.

Example-3

41FF STA 526AH



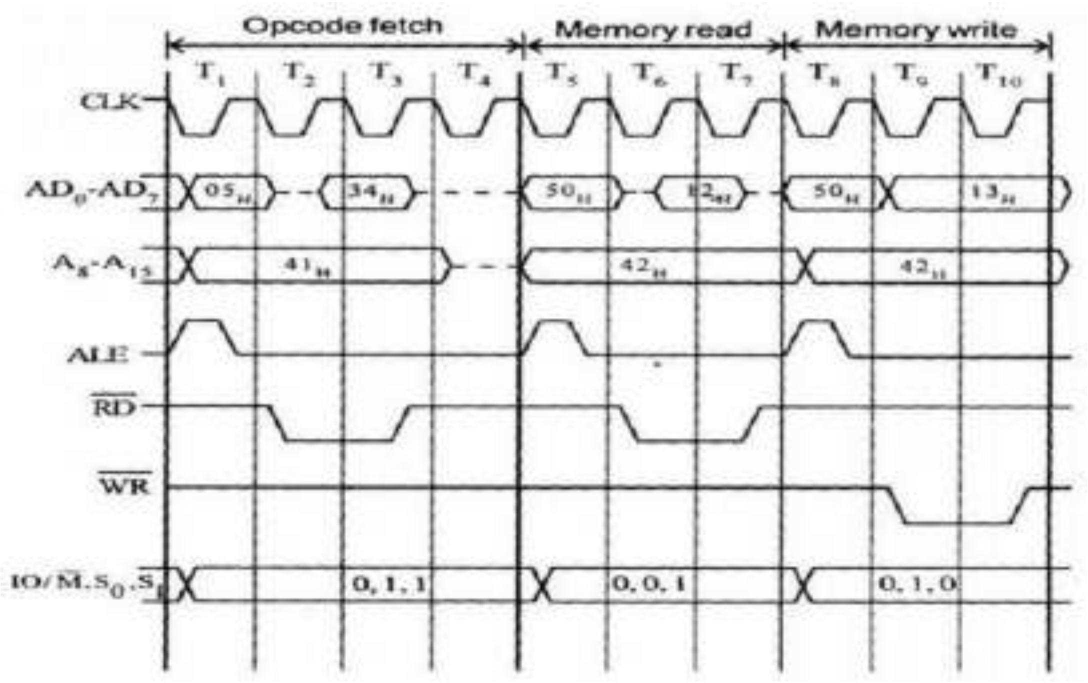
 STA means Store Accumulator -The contents of the accumulator is stored in the specified address (526A).

- The opcode of the STA instruction is said to be 32H. It is fetched from the memory 41FFH
- Then the lower order memory address is read (6A). - Memory Read Machine Cycle
- Read the higher order memory address (52).- Memory Read Machine Cycle
- The combination of both the addresses are considered and the content from accumulator is written in 526A. - Memory Write Machine Cycle
- Assume the memory address for the instruction and let the content of accumulator is C7H. So, C7H from accumulator is now stored in 526A.

Example-4

4105 INR M

- Fetching the Opcode 34H from the memory 4105H. (OF cycle)
- Let the memory address (M) be 4250H. (MR cycle -To read Memory address and data)
   Let the content of that memory is 12H.
- Increment the memory content from 12H to 13H. (MW machine cycle)



Counter and time delay

When the delay subroutine is executed, the microprocessor does not execute other tasks. For the delay we are using the instruction execution times. executing some instructions in a loop, the delay is generated. There are some methods of generating delays. These methods are as follows.

- Using NOP instructions
- Using 8-bit register as counter
- Using 16-bit register pair as counter. **Using NOT instructions:**
- One of the main usage of NOP instruction is in delay generation.
- The NOP instruction is taking four clock pulses to be fetching, decoding and executing.
- In the 8085 MPU the internal clock frequency is 3MHz.
- So, from that we can easily determine that each clock period is 1/3 of a microsecond.
- So, the NOP will be executed in 1/3 * 4 = 1.333µs. **Using 8-bit register as counter:**
- Counter is another approach to generate a time delay.
- In this case the program size is smaller.
- So, in this approach we can generate more time delay in less space.
- The following program will demonstrate the time delay using 8-bit counter. MVI B,FFH

```
LOOP: DCR B
    JNZ LOOP
    RET
```

- Here the first instruction will be executed once, it will take 7 T-states.
- DCR C instruction takes 4 T-states.
- This will be executed 255 (FF) times.
- The JNZ instruction takes 10 T-states when it jumps (It jumps 254 times), otherwise it will take 7 T-States.
- And the RET instruction takes 10 T-States.
  7 + ((4*255) + (10*254)) + 7 + 10 = 3584.
- So, the time delay will be 3584 * 1/3µs = 1194.66µs.

- So, when we need some small delay, then we can use this technique with some other values in the place of FF.

This technique can also be done using some nested loops to get larger delays. The following code is showing how we can get some delay with one loop into some other loops.

```
MVI  B,FFH
L1: MVI C,FFH
L2: DCR C
    JNZ L2
    DCR  B
    JNZ L1
    RET
```

From this block, if we calculate the delay, it will be nearly 305µs delay. It extends the time of delay.

Using 16-bit register-pair as counter:

- Instead of using 8-bit counter, we can do that kind of task using 16-bit register pair.
- Using this method more time delay can be generated.
- This method can be used to get more than 0.5 seconds delay.

Program	Time (T-States)
LXI B,FFFFH **LOOP: DCX B** **MOV A,B** **ORA C** **JNZ LOOP** **RET**	10 6 4 4 10 (For Jump), 7(Skip) 10

From that table, if we calculate the time delay:

10 + (6 + 4 + 4 + 10) * 65535H – 3 + 10 = 17 + 24 * 65535H = 1572857.
So, the time delay will be 1572857 * 1/3µs = 0.52428s. Here we are getting nearly 0.5s delay.

Assembly language program

Example-1

Write an assembly language program to add two 8-bit numbers 45H and 32H in 8085 Microprocessor and store the result in 2050H. The starting address of the program is taken as 2000.

Program address	Mnemonics	Operands	comments
2000	MVI	A,45	Load 1st data 45H in ACC
2002	MVI	B,32	Load 2nd data 32H in B
2004	ADD	B	A+B=A
2005	STA	2050	Store the result in 2050H
2008	HLT		Stop the program

O/P address **Result**

2050H 77H

Example-2

Write an ALP to add 2 8-bit numbers stored in memory location 2050H and 2051H. Result can be 8/16 bit and store it in 2052H and 2053H.

Program address	**Label**	**Mnemonics**	**Operands**	**Comments**
2000		MVI	C,00	Initialize the carry
2002		LXI	H,2050	Get the 1st data
2005		MOV	A,M	Load 1st data in ACC
2006		INX	H	Get 2nd data
2007		ADD	M	Add both data
2008		JNC	LOOP	If no carry, jump to LOOP
200B		INR	C	If carry, increment register C
200C	LOOP	STA	2052	Store the sum in 2052
2010		MOV	A,C	Move carry to ACC
2011		STA	2053	Store carry in 5053
2014		HLT		Stop the program

Without carry

I/P address	**Data**
2050	53
2051	27

With carry

I/P address	**Data**
2050	D9
2051	62

O/P address	Result
2052	7A
2053	00

O/P address	Result
2053	3B
2053	01

Basic Interfacing concept

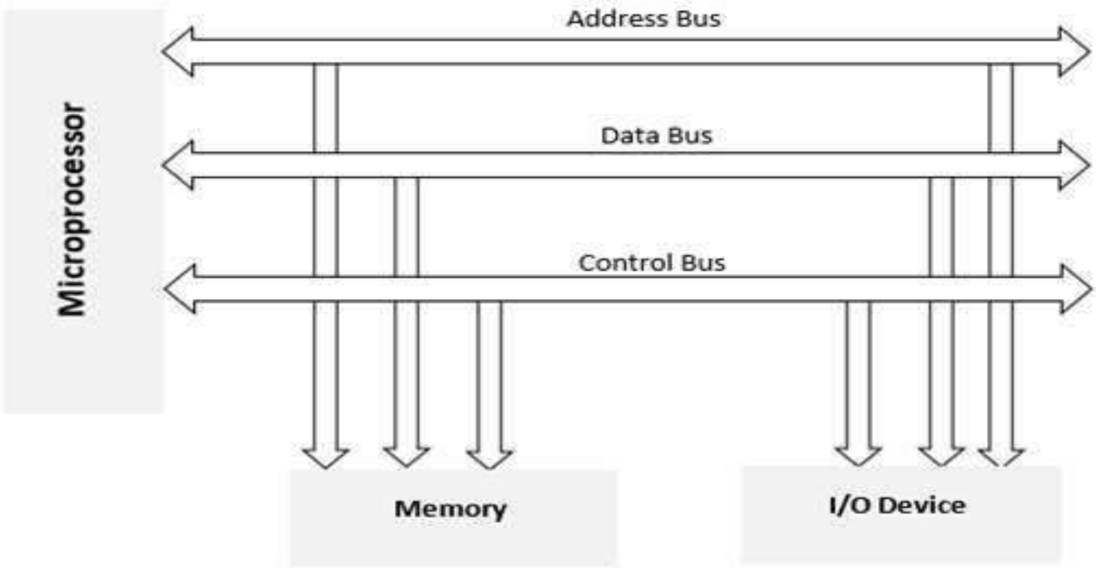
Interface is the path for communication between two components. Interfacing is of two types, memory interfacing and I/O interfacing.

Memory interfacing

- Memory interfacing is used to provide more memory space to accommodate complex programs for more complicated systems.
- Types of memories which are most commonly used to interface with 8085 are RAM, ROM, and EEPROM.
- 8085 can access 64kB of external memory.
- It can be explained as- total number of address lines in 8085 are 16, therefore it can access 2^16 = 65535 locations i.e., 64kB

I/O interfacing

- There are various communication devices like the keyboard, mouse, printer, etc.
- So, we need to interface the keyboard and other devices with the microprocessor by using latches and buffers.
- This type of interfacing is known as I/O interfacing. **Block diagram of memory and I/O interfacing**



Memory mapped I/O and I/O mapped I/O In
Memory Mapped Input Output −

- We allocate a memory address to an Input-Output device.
- Any instructions related to memory can be accessed by this Input-Output device. The Input-Output device data are also given to the Arithmetic Logical Unit.

Input-Output Mapped Input Output −

- We give an Input-Output address to an Input-Output device
- Only IN and OUT instructions are accessed by such devices.

- The ALU operations are not directly applicable to such Input-Output data. So as a summary we can mention that −
- I/O is any general-purpose port used by processor/controller to handle peripherals connected to it.
- I/O mapped I/Os have a separate address space from the memory. So, total addressed capacity is the number of I/Os connected and a memory connected. Separate I/Orelated instructions are used to access I/Os. A separate signal is used for addressing an I/O device.
- Memory-mapped I/Os share the memory space with external memory. So, total addressed capacity is memory connected only. This is underutilisation of resources if your processor supports I/O-mapped I/O. In this case, instructions used to access I/Os are the same as that used for memory.
- Let's take an example of the 8085 processor. It has 16 address lines i.e., addressing capacity of 64 KB memory. It supports I/O-mapped I/Os. It can address up to 256 I/Os.
- If we connect I/Os to it an I/O-mapped I/O then, it can address 256 I/Os + 64 KB memory. And special instructions IN and OUT are used to access the peripherals. Here we fully utilize the addressing capacity of the processor.
- If the peripherals are connected in memory mapped fashion, then total devices it can address is only 64K. This is underutilisation of the resource. And only memoryaccessing instructions like MVI, MOV, LOAD, SAVE are used to access the I/O devices.

UNIT-5 Interfacing and support chips

8255 Programable Peripheral Interface (PPI)

- PPI 8255 is a general purpose programmable I/O device designed to interface the CPU with its outside world such as ADC, DAC, keyboard etc.
- We can program it according to the given condition. It can be used with almost any microprocessor.
- It consists of three 8-bit bidirectional I/O ports (24 I/O lines) which can be configured as per the requirement.

Ports of 8255A

8255A has three ports, i.e., PORT A, PORT B, and PORT C.

- Port A (PA0-PA7) contains one 8-bit output latch/buffer and one 8-bit input buffer.
- Port B (PB0-PB7) is similar to PORT A.
- Port C can be split into two parts, i.e., PORT C lower (PC0-PC3) and PORT C upper (PC7-PC4) by the control word.

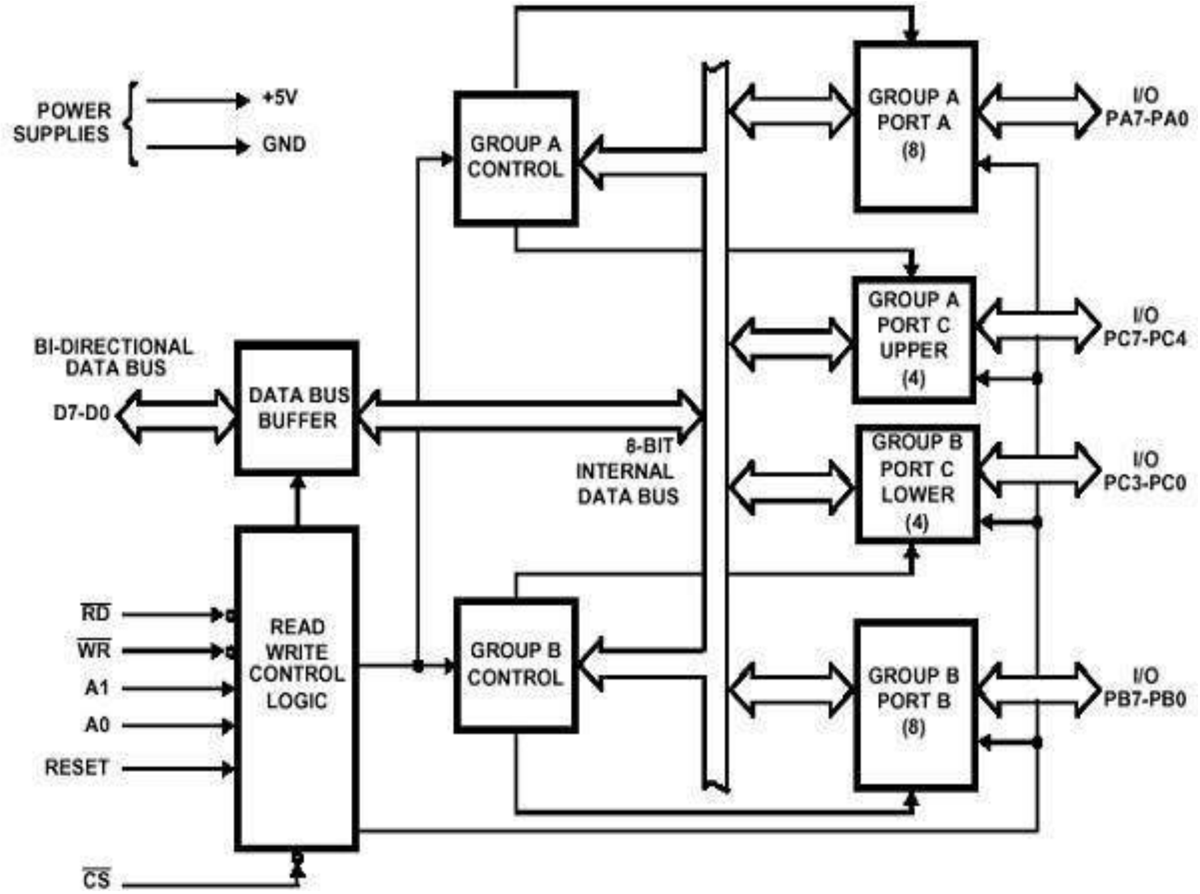
These three ports are further divided into two groups, i.e., Group A includes PORT A and upper PORT C. Group B includes PORT B and lower PORT C. These two groups can be programmed in three different modes, i.e., the first mode is named as mode 0, the second mode is named as Mode 1 and the third mode is named as Mode 2.

Features of 8255A

The prominent features of 8255A are as follows − It consists of 3 8-bit I/O ports i.e., PA, PB, and PC.

- Address/data bus must be externally demultiplexed.
- It is TTL compatible.
- It has improved DC driving capability.

8255 Architecture



Control group A

Control group A consist of port A and port C upper.

Control group B

Control group B consists of port C lower and port B.

Data Bus Buffer

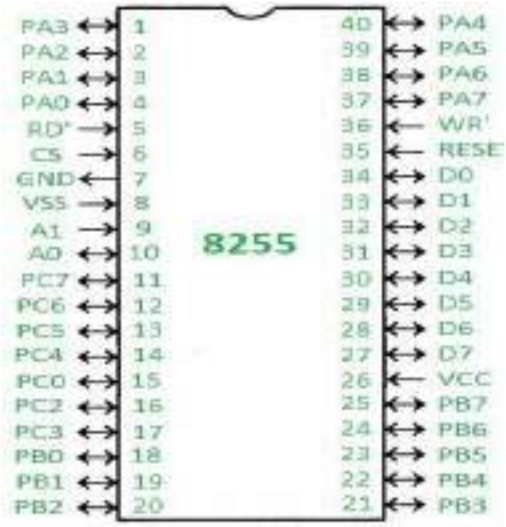
- It is a tri-state 8-bit buffer, which is used to interface the microprocessor to the system data bus.
- Data is transmitted or received by the buffer as per the instructions by the CPU.
- Control words and status information is also transferred using this bus.

Read/Write Control Logic

- This block is responsible for controlling the internal/external transfer of data/control/status word.
- It accepts the input from the CPU address and control buses, and in turn issues command to both the control groups.
- Depending upon the value if CS', A1 and A0 we can select different ports in different modes as input-output function or BSR.
- This is done by writing a suitable word in control register (control word D0-D7).

CS’	A1	A0	Selection
0	0	0	PORT A
0	0	1	PORT B
0	1	0	PORT C
0	1	1	Control Register
1	X	X	No Selection

Pin diagram



CS
It stands for Chip Select. A LOW on this input selects the chip and enables the communication between the 8255A and the CPU. It is connected to the decoded address, and A0 & A1 are connected to the microprocessor address lines.
WR
It stands for write. This control signal enables the write operation. When this signal goes low, the microprocessor writes into a selected I/O port or control register.
RESET
This is an active high signal. It clears the control register and sets all ports in the input mode
RD
It stands for Read. This control signal enables the Read operation. When the signal is low, the microprocessor reads the data from the selected I/O port of the 8255.
A0 and A1
These input signals work with RD, WR, and one of the control signals. Following is the table showing their various signals with their result.

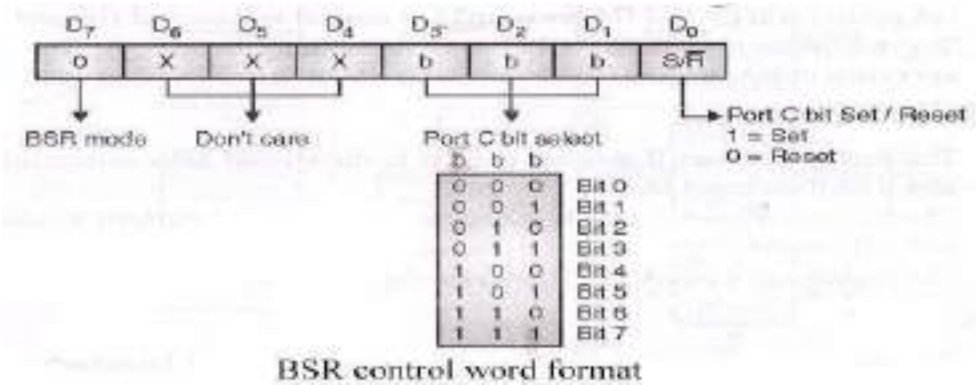
A_1	A_0	RD	WR	CS	Result
0	0	0	1	0	**Input Operation** PORT A → Data Bus
0	1	0	1	0	PORT B → Data Bus
1	0	0	1	0	PORT C → Data Bus
0	0	1	0	0	**Output Operation** Data Bus → PORT A
0	1	1	0	0	Data Bus → PORT A

1	0	1	0	0	Data Bus → PORT B

Operating Modes

1. BSR (bit set-reset) mode-

If MSB of control word (D7) is 0, PPI works in BSR mode. In this mode only port C bits are used for set or reset.



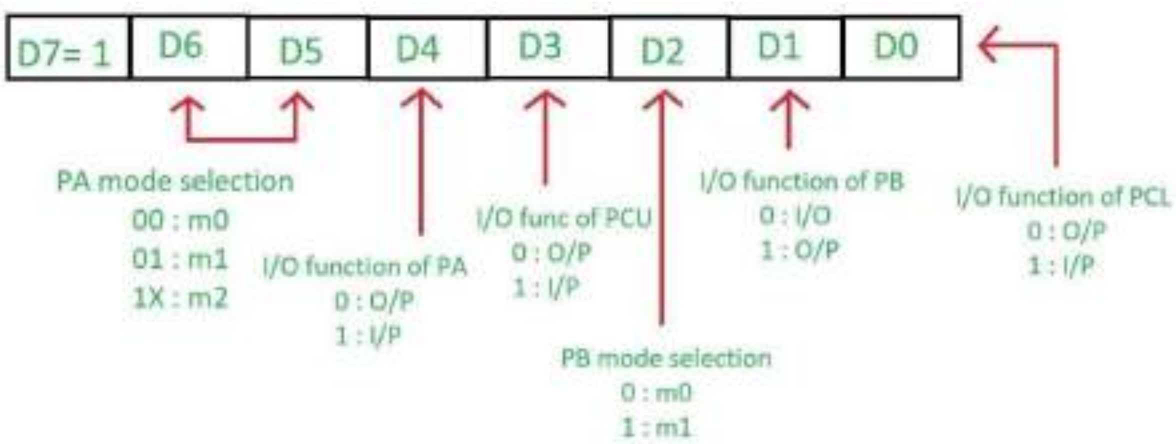
BSR control word format

2. I/O mode-

Mode 0 − In this mode, Port A and B is used as two 8-bit ports and Port C as two 4-bit ports. Each port can be programmed in either input mode or output mode where outputs are latched and inputs are not latched. Ports do not have interrupt capability.

Mode 1 − In this mode, Port A and B is used as 8-bit I/O ports. They can be configured as either input or output ports. Each port uses three lines from port C as handshake signals. Inputs and outputs are latched.

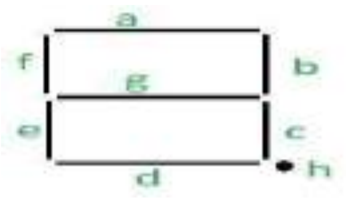
Mode 2 − In this mode, Port A can be configured as the bidirectional port and Port B either in Mode 0 or Mode 1. Port A uses five signals from Port C as handshake signals for data transfer. The remaining three signals from Port C can be used either as simple I/O or as handshake for port B.



Seven segment LED display

A seven-segment LED is a kind of LED (Light Emitting Diode) consisting of 7 small LEDs it usually comes with the microprocessor's as we commonly need to interface them with microprocessors like 8085.

Structure of Seven Segments LED:



- It can be used to represent numbers from 0 to 8 with a decimal point.
- We have eight segments in a Seven Segment LED display consisting of 7 segments which include '.'.
- The seven segments are denoted as "a, b, c, d, e, f, g, h" respectively, and '.' is represented by "h".

Interfacing Seven Segment Display with 8085:

We will see a program to Interfacing Seven Segment Display with 8085 using 8255.
Note logic needed for activation –
Common Anode – 0 will make an LED glow.
Common Cathode – 1 will make an LED glow.
Common Anode Method:
Here we are using a common anode display therefore 0 logic is needed to activate the segment. Suppose to display number 9 at the seven-segment display, therefore the segments F, G, B, A, C, and D have to be activated.
The instructions to execute it is given as,

```
    MVI  A,99
OUT 00
```

- First, we are storing the 99H in the accumulator i.e., 10010000 by using MVI instruction.
- By OUT instruction we are sending the data stored in the accumulator to the port 00H. Common Cathode Method:

Here we are using common cathode 1 logic is needed to activate the signal. Suppose to display number 9 at the seven-segment display, therefore the segments F, G, B, A, C, and D have to be activated.
The instructions to execute it is given as,

```
    MVI  A,6F
OUT 00
```

- First, we are storing the 6FH in the accumulator i.e., 01101111 by using MVI instruction.
- By OUT instruction we are sending the data stored in the accumulator to the port 00H.

Traffic light controller

The traffic lights are interfaced to Microprocessor system through buffer and ports of programmable peripheral Interface 8255. So the traffic lights can be automatically switched ON/OFF in desired sequence. The Interface board has been designed to work with parallel port of Microprocessor system.

Working Program

Design of a microprocessor system to control traffic lights. The traffic should be controlled in the following manner.
1) Allow traffic from W to E and E to W transition for 20 seconds.
2) Give transition period of 5 seconds (Yellow bulbs ON)
3) Allow traffic from N to 5 and 5 to N for 20 seconds
4) Give transition period of 5 seconds (Yellow bulbs ON) 5) Repeat the process.

Source Program:

```
         MVI A, 80H:          Initialize 8255, port A and port B
          OUT 83H (CR):       in output mode
START: MVI A, 09H
         OUT    80H    (PA):  Send data on PA to glow R1 and R2
MVI A, 24H
         OUT 81H (PB):                 Send data on PB to glow G3 and G4
         MVI C, 28H:          Load multiplier count (40ıo) for delay
         CALL DELAY:                 Call delay subroutine
         MVI A, 12H

        OUT (81H) PA:         Send data on Port A to glow Y1 and Y2
        OUT (81H) PB:         Send data on port B to glow Y3 and Y4
        MVI C, 0AH:           Load multiplier count (10ıo) for delay
        CALL: DELAY:                 Call delay subroutine
        MVI A, 24H
        OUT (80H) PA:  MVI    Send data on port A to glow G1 and G2
        A, 09H
        OUT (81H) PB:         Send data on port B to glow R3 and R4
        MVI C, 28H:           Load multiplier count (40ıo) for delay
        CALL DELAY:  MVI             Call delay subroutine
        A, 12H
        OUT PA:               Send data on port A to glow Y1 and Y2
        OUT PB:               Send data on port B to glow Y3 and Y4
        MVI C, 0AH:           Load multiplier count (10ıo) for delay
        CALL DELAY:                  Call delay subroutine
        JMP START
Delay Subroutine:
DELAY:  LXI D, Count:                 Load count to give 0.5 sec delay
BACK:  DCX D:  MOV A, D              Decrement counter

        ORA E:                       Check whether count is 0
        JNZ BACK:             If not zero, repeat
        DCR C:                Check if multiplier zero, otherwise repeat
        JNZ DELAY
        RET:                  Return to main program
```

Square wave generator

- With 00H as i/p to DAC, analog o/p is -5V, and with FFH as i/p, analog o/p is +5V.
- I/P 00H and FFH at regular intervals generate square wave. The frequency can be varied by varying the time delay.

Algorithm

Initialize the control word of 8255 to operate in I/O mode for port A and B & C to operate in o/p mode.

Program

```
        MVI A,80
        OUT CWR      initialize the control word
LOOP: MVI A,00
```

```
        OUT PA
        CALL DELAY
        MVI A,FF
        OUT PA
        CALL DELAY
        JMP LOOP
DELAY: MVI C,85
BACK: DCR C
        JNZ BACK
        RET
```

CONDUCTING MATERIAL
1.1-1.11

INTRODUCTION: -

Materials which are used commercially for conducting electricity can be classed as conducting materials and materials which are used for preventing the flow of electricity can be classed as non-conducting or insulating materials.

Conducting materials have very low values of resistivity as compared to insulating material. To determine the extent to which a material has conducting or insulating property, we should know the value of its resistivity. Conducting materials can further be subdivided into low resistivity and high re-
resistivity materials.

Resistivity and factors affecting resistivity:

Resistivity: The reader already knows Ohm's law which can be written as V =IR, where, V is the value of the voltage between the terminals of a current carrying conductor, I is the current flowing through the conductor and R is the resistance of the conductor. The resistance R of any given material is directly proportional to its length 'l and inversely proportional to its cross-sectional area ‘a'

Thus, $R \text{ a } \frac{l}{a}$ or $R = \rho \frac{l}{a}$ ohm, ...(2.1)

Where, ‘ρ’ is the coefficient of proportionality is called the resistivity or specific resistance of the material. In expression (2.1),

R = resistance of the material in ohm

r = resistivity of the material in ohm-m

a = area of cross=section of the material in sq-m

l = length of the material in meters.

Effect of Temperature on Resistivity: The most important factor which affects the value of resistivity is the temperature The resistance of most of the conducting material increases with temperature. The change in resistance of a material per ohm per degree change in temperature is called the "temperature coefficient of resistance of that material.

$R_t = R_0\,(1+\text{a}\ t)$ ....(2.2)

Where, R_t and R_0 are respectively the resistances of the conductor at t degree and zero degree centigrade and a the temperature coefficient of resistance.

If the resistance of the same material at any other temperature t_1 degree centigrade be R_{t1} then according to expression (2.2):

$$R_{t1} = R_0\,(1+\text{a}\ t_1\,)$$

Dividing expression (2.3) by expression (2.4) we ger,

$$\frac{R_{t1}}{R_t} = \frac{1+\alpha\,t_1}{1+\alpha\,t} = \frac{1+\alpha\,t+\alpha\,t_1-\alpha\,t}{1+\alpha\,t}$$

(Adding and subtracting α t in numerator)

$$= \frac{1+\alpha t}{1+\alpha t} = \frac{\alpha\,(\,t_1-\,t)}{1+\alpha\,t}$$

$$= 1+\frac{\alpha\,(\,t_1-t)}{1+\alpha\,t}$$

$$R_{t1} = R_t\ 1+[\frac{\alpha}{1+\alpha\,t}\,(\,t_1-\,t)] \qquad ....(2.4)$$

This means that the resistance at any temperature t_1 degrees can be calculated if the resistance at t degrees is known.

The resistance of a conductor changes with temperature according to the law Expression (2.4) is very important. It enables the designer of electrical equipment to determine by calculation the 12 R losses in the windings of equipment like motors and transformers. For this, he must know the resistance of the winding at the operating temperature. Assuming that the operating temperature of a transformer winding is 65 degrees centigrade above the ambient temperature (say of 3) degrees centigrade, then = 30 degrees C and = 95 degrees C. The value of resistance R, at ambient temperature (Let at 30 degrees C) is generally known to the designer. He can find from data books the value of the temperature coefficient and then calculate the resistance R. At the operating temperature (i.e., at 95 degrees C) by applying expression (2.4)-The relationship between temperature coefficient of resistance with change in temperature can also be found out by following ways. Let us assume that R. R, and R be the resistance of a conductor at 12 respectively then,

$$R_2 = R_1[1+\alpha_1\,(\,t_2-\,t_1)] \qquad ....(2.5)$$

$$R_3 = R_1[1+\alpha_1\,(\,t_3-\,t_1)] \qquad .....(2.6)$$

$$= R_1[1+\alpha_1\,(\,t_3-\,t_1)] \qquad ......(2.7)$$

Or,

$$\frac{R3}{R_2} = [1+\alpha_2\,(\,t_3-\,t_2)] \qquad ......(2.8)$$

Dividing equation (2.6) by (2.5), we get,

$$\frac{R_3}{R_2} = \frac{[1+\alpha_1\,(\,t_3-\,t_1)]}{[1+\alpha_1\,(\,t_2-\,t_1)]}$$

$$= \frac{1 + \alpha_1 (t_2 - t_1) + \alpha_1 (t_3 - t_1) - \alpha_1 (t_2 - t_1)}{1 + \alpha_1 (t_2 - t_1)}$$

$$= 1 + \frac{\alpha_1 (t_3 - t_2)}{1 + \alpha_1 (t_2 - t_1)}$$

$$= 1 + \frac{\alpha_1}{1+\alpha_1 (t_2 - t_1)} \times (t_3 - t_2) \qquad .....(2.9)$$

Comparing equation (2.8) and (2.9) we get,

$$= 1 + \alpha_2 (t_3 - t_2) = 1 + \frac{\alpha_1}{1+\alpha_1 (t_2 - t_1)} \times (t_3 - t_2)$$

Or $\qquad \alpha_2 = \frac{\alpha_1}{1+\alpha_1 (t_2 - t_1)}$

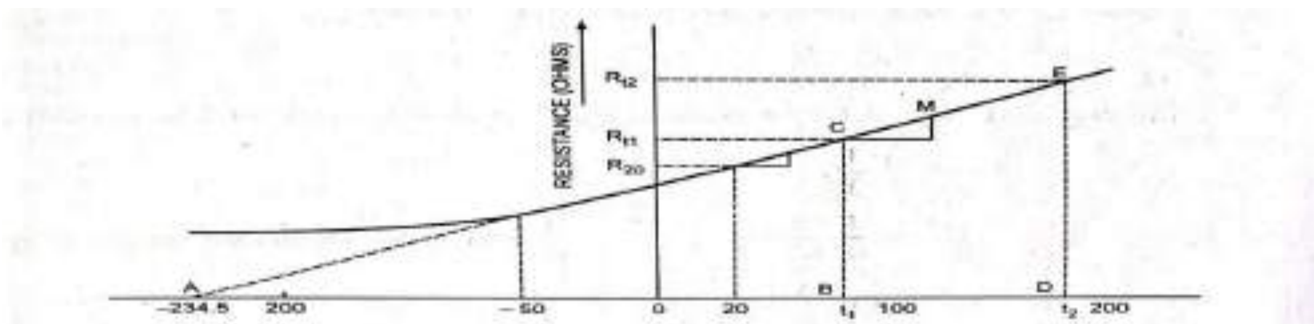
Or $\qquad \alpha_2 = \frac{\alpha_1}{\frac{1}{\alpha_1} (t_2 - t_1)}$

The test engineer can determine the actual temperature rise of the winding in this way: The resistance of the winding R_t, when the machine is not connected to the supply and when its winding is at an ambient temperature of t degrees, is measured. Then the equation is loaded under normal operating conditions for a sufficiently long time. Then its hot resistance R is measured immediately after switching off the supply. By knowing the value of a for the winding material, the temperature rise (4-t) can be calculated by applying expression (2.4).

It is important to note that the temperature coefficient of resistance, a is not constant. For understanding this let us rewrite expression (2.2) thus:

R = R, (1 + at) or R = R, + Ro at. (2.10)

This can be expressed graphically as in figure 2.1. This figure is drawn for annealed copper in which case resistance becomes zero at - 234.5°C (This value has been arrived at by extrapolating the graph; but in actual fact the graph tends to deviate from straight line at about-50°C as shown in figure 2.1). It is seen that the change in resistance with change in temperature is constant but since the graph does not pass through the origin the resistance is not directly proportional to the temperature. This is in fact also obvious from expression (2.5).



The Variation of Resistance with Temperature for Annealed Copper
Fig. 2.1.

Since the temperature coefficient has been defined as the change in resistance per ohm per degree change in temperature, it means that:

at $20°C$, $\alpha_{20} = \frac{\Delta R}{R_{20}}$;

at $0°C$, $\alpha_0 = \frac{\Delta R}{R_0}$;

Where, ΔR= change on resistance for 1° change in temperature,

R_0= resistance at $0°C$,

R_{20}= resistance at $20°C$,

In general :

$$\alpha_t = \frac{\Delta R}{R_t}$$

Where, $R_t = resistance\ at\ t$°C

a$_t$= $temperature\ coefficient\ of\ resistance\ at\ t$°C

$Thus\ for\ temperature\ t_1\ and\ t_2$, we have:

$$\alpha_{t1} = \frac{\Delta R}{R_{t1}} \qquad ....(2.11)$$

and

$$\alpha_{t2} = \frac{\Delta R}{R_{t2}} \qquad ....(2.12)$$

Consider the similar triangle ADE and ABC, we have:

$$\frac{DE}{BC} = \frac{AD}{AB} \quad OR \quad \frac{R_{t2}}{R_{t1}} = \frac{234.5 + t_2}{234.5 + t_1}$$

The equation is true for anealed copper only. In general the straighy line graph will meet the temperature axis at different point depending upon the material. Let the point of intersection in general be taken as β degree below zero degree C. We may rewrite ecpression (2.13) thus:

$$\frac{R_{t2}}{R_{t1}} = \frac{\beta + t_2}{\beta + t_1}$$

Consider, in figure 2.1, the similar right angled triangles CLM and ABC, we have:

$$\frac{LM}{BC} = \frac{CL}{AB}$$

Or $\frac{\Delta R}{R_{t1}} = \frac{1}{234.5 + t_1}$ (True for annealed copper only)

We already knoe thet $\frac{\Delta R}{R_{t1}} = \alpha_{t1}$ (See expression 2.11)

Therefore, $\alpha_{t1} = \frac{1}{234.5 + t1}$ ...(2.15)

Expression (2.15) is true for annealed copper only. In ganeral for any material :

$$\alpha_t = \frac{1}{\beta + t1}$$

If β is known, the temperature coefficient at any temperature can be found for any particular material.

Table 2.1 gives comparative study of the value of resistivity, temperature coefficient, density and melting point of different materials. Depending up on the application any of these materials may be selected.

Table. 2.1. Values of resistivity, temperature coefficient, density and melting point for different materials.

Material	Resistivity (ohm-m) at 20°C × 10^{-8}*	Temperature coefficient per degree C at 20°C × 10^{-4}*	Density	Melting point (degrees C)
Aluminium (cast soft)	2.8	35	2.68	655
Aluminium (hard drawn)	2.9	35	2.71	630
Carbon	400 to 1200	− 12 to − 60	1.9 to 2.3	3450
Copper (annealed)	1.72	39	8.89	1084
Copper (hard drawn)	1.77	39	8.89	1084
Iron (cast)	75 to 98	–	7.80	1500 to 1530
Lead	21	41	11.40	327
Nickel (commercial)	10.5	40	8.85	1450
Silver	1.60	40	10.50	960
Tin	11.50	46	7.30	232
Tungsten	5.50	50	18.80	3300
Silicon steel	50 to 60	–	7.70	–
Carbon steel (high carbon)	15 to 45	2 to 40	–	–
Carbon steel (up to 0.4 % carbon)	10 to 14	40 to 50	7.80	1350
Nichrome	100	4.4	8.15	1538
Brass	7	15 to 20	8.40 to 8.70	–
Manganin	48	0.5	8.40	102
Constantan	52	0.25 to 0.5	8.90	–

* 10^{-8} and 10^{-4} are factors by which the values in the corresponding columns are to be multiplied.

Example 2.1. A coil of a relay is made of copper wire. At a temperature of 20°C, the resistance of the coil is 400 ohms. Calculate the resistance of the coil at a temperature of 80°C. Temperature coefficient of copper is 0.0038 ohm per degree C at 0°C.

Solution : We know the relation :

$$R_{t_1} = R_t\left[1+\frac{\alpha}{1+\alpha t}(t_1-t)\right] \qquad \text{(from expression 2.4)}$$

Given, R_{20} = 400 ohms, α = 0.0038 ohm per degree C at zero degree C. It is required to find R_{80}.

Here, $R_{t_1} = R_{80}$, $R_t = R_{20}$.

Putting the given data in the above expression we get,

$$R_{80} = R_{20}\left[1+\frac{\alpha}{1+\alpha\times 20}(80-20)\right]$$

$$= 400\left[1+\frac{0.0038\times 60}{1+0.0038\times 20}\right]$$

$\therefore$ $R_{80} = 484$ ohms.

Example 2.2. Calculate the resistance of a wire at 50°C which is 300 m long and has an area of cross-section of 25 mm². The wire is made of aluminium. Resistivity of aluminium at 15°C is 2.78 ohm-m. Temperature coefficient of aluminium is 0.004 ohm/degree C at 0°C.

Solution : Given $l = 300$ m, $\alpha = 25\ mm^2$, ρ at 15°C = 2.78 ohm-m.

$$R_{15} = \rho_{15}\frac{1}{\alpha} = 2.78\times\frac{300}{25\times 10^{-6}} = 33.4\times 10^6 \text{ ohms.}$$

$$= 33.4\ M\ \Omega$$

$$R_{50} = R_{15}\left[1+\frac{\alpha}{1+\alpha\times 15}(50-15)\right]$$

$$= 33.4\left(1+\frac{0.004\times 35}{1+0.004\times 15}\right)$$

or $R_{50} = 33.4\ (1 + 0.132) = 38\ M\Omega$

Example 2.3. A coil is made of copper-wire. At 15°C the resistance of the coil is 250 ohms. What will be the temperature of the same coil if its resistance is 300 ohms ?

Solution : Let R_{tx} be the resistance (= 300 ohm) at t_x degree C

Then, $$R_{tx} = R_{15}\left[1+\frac{\alpha}{1+\alpha\times 15}(t_x-15)\right]$$

or $$300 = 250\left[1+\frac{0.0038}{1+0.0038\times 15}(t_x-15)\right]$$

From which $t_x = 67$°C.

Effect of alloying on resistivity: Alloying is another factor which affects the resistivity of a material. By adding some impurities, a small percentage of some other material) to a metal its resistivity can be increased Alloys have higher resistivity than the pure base metal. At the same time, when a metal is alloyed, it also acquires properties like higher mechanical strength which are needed for certain applications. For example, when copper is alloyed with zinc, the alloyed material is called brass (60 % copper, 40 %zinc). By alloying copper with zinc its resistivity is increased i.e. conductivity is decreased by about 4 times (see table 2.1). But the tensile strength of brass is much more than that of copper and therefore may be used for making structural products such as rods, shafts, heavy plates, plug point, socket outlets, knife switches etc. where high strength and hardness are usually desirable.

Effect of mechanical stressing on resistivity: The resistivity of a material also changes under the influence of mechanical treatment The fabrication of conductor from the ingot to the final stage comprises initially hot working and finally cold-drawing Cold-working operation (stressing) distorts the crystal structure of the metal. This generally tends to harden the material increase its tensile strength and increase slightly its resistivity (see table 2.1) The increase in tensile strength in very useful for many purposes such as overhead conductor. That is why many types of conductors are finally drawn in cold stage in which case they are identified as hard drawn Although mechanical stressing increases the resistivity i.e. decreases the conductivity, annealing (Heat treatment process) restores the electrical conductivity by establishing regularity in crystal structure.

Classification of conducting materials into low-resistivity and high resistivity materials:

Low resistivity material: - Low resistivity materials are used in house wiring, as conductors for power transmission and distribution, in the windings of transformer and machines like motors and generators. In fact, low resistivity materials are used in all such applications where power loss and voltage drop should be low. Copper and aluminium are examples of commercially acceptable low resistivity material. Silver has lower resistivity than copper but because of its prohibitive cost its use commercially as a conductor is not feasible.

A low resistivity material, besides possessing low value of resistivity should also possess the following additional properties for rent against each:

(a) Low temperature coefficient: This means that the change of resistance with change in temperature should be low. This is necessary to avoid variation in voltage drop and power loss with changes in temperature. For example, the resistance of transmission lines which are very long will increase when ex-posed to hot summer sun. This will cause increase in voltage drop and power loss in the transmission line. The windings of electrical machines and apparatus become hot when loaded. This causes temperature rises and if the conducting material of the winding has high temperature coefficient of resistance, the voltage drops and power loss in the winding will be high.

(b) Sufficient mechanical strength: The overhead line conductors used for transmission and distribution of electrical power are subject to stresses due to wind and their own weight. The conducting materials used for the windings of transformers, motors and generators develop mechanical forces when loaded which can become very large if a high current flows due to a short circuit Also, when the coils for the windings of such equipment are made on former the conducting material is subject to mechanical stresses. Therefore, to with stand the mechanical stress, developed in the above mentioned application the conducting material should possess sufficient mechanical strength.

c) Ductility: - Ductility is that property of a material which allows it to be drawn out into a wire. Conductors are required in different sizes and shapes In some applications round wire section is used, while in others rectangular wire section is used The conducting material should be ductile enough to enable itself being drawn into different sizes and shapes.

(d) Solderability: Conductors have often to be jointed. The joint should offer minimum contact resistance. A simple joint would be to twist the conductor with the material to which it is to be jointed. But this gives high contact resistance. Minimum contact resistance results if the joint is soldered A material do not tend themselves to proper soldering. So while selecting conducting material, this point should be kept in view

(e) Resistance to corrosion: The conducting material should be such the it is not corroded when used in out-door atmosphere. Note 1 The reader should not lead himself to believe that all conducting materials should possess all the above mentioned properties. Depending upon the applications an appropriate material should be chosen which may not have all the above properties but those which the particular applications called for.

High resistivity materials:- High resistivity materials are used for making resistance elements for heating devices, starters for electric motor resistance used in precision measuring instruments, loading resistances and rheostats and filaments for incandescent lamps, In fact, high resistivity materials are used in all such applications where a large value of resistance is required If low resistivity materials were used for such application the length of the wire would be too large which would increase to a large extent the of the equipment. A high resistivity material besides possessing high values of resistivity should also posses the following additional properties for reasons mentioned against each.

(a). **Low temperature coefficient:-** High resistivity material are often used as shunts in electrical measuring instruments in making wire-wound precision resistance and resistance boxes. For Such precision application an important requirement is that the material of the element should have negligible temperature coefficient of resistance as otherwise the accuracy of measurement will be reduced.

(b). **High melting point:-** In applications like loading rheostats and starters for electrical motors the material of the resistance element should be able to withstand high temperature for a long time without melting. The Consideration of the resistance" temperature coefficient in these cases is also important but comparatively high values than those mentioned in above are permissible.

The consideration of high melting point is important also for resistance material used in electrical heating devices like room heaters, furnace, etc.

(c). **No tendency for oxidation:-** Material used as high resistance elements in heating appliances should be able to withstand high temperature for a long time without oxidation. This is because of an oxide layer is formed on the heating element the amount of heat radiation will reduce.

(d). **Ductility:-** High resistance material are required in the shape of very thin wires in the case of precision wire - wound resistors and in the shape of thick wires in case of the elements used in ovens, heaters, starters, etc. High resistance materials to be used for such application should therefore be capable of being drawn in to wires of different sizes and further be capable of being coiled.

(d). **High mechanical strength:-** High resistivity materials to be used for applications where the wire must be very thin and required to have high tensile strength as otherwise they may break during the braw of the wire or during the assembly and subsequent operation.

Example 2.4 A heater element 18 made of nichrome wire having resistivity equal to 100 x 10 ohm metre. The diameter of the wire is 0.4 mm
(a) Calculate the length of the wire required to get a resistance of 40 ohms and 1000 watts
(b) Calculate also the length required if the material of the element was copper having same cross-sectional area. Assume resistivity of copper to be 1.732 x 10- ohm-metre.

Solution: (a) When the element material is nichrome
ρ = 100 x 10 ohm-m.
d = 0.4 x 10 mm
a=12.6 x 10m
R= 40 ohms

We know that:

$$R = \rho \frac{l}{a}$$

Therefore, $$40 = 100 \times 10^{-8} \times \frac{l}{12.6 \times 10^{-8}}$$

$$l = 5\, metres$$

Or

(b) When the element material is copper ,

$$\rho = 1.72 \times 10^{-8}\, ohm - metre$$

$$a = 12.6 \times 10^{-8}\, m^2$$

R= 40 ohms.

Therefore, $$40 = 1.72 \times 10^{-8} \times \frac{l}{12.6 \times 10^{-8}}$$

$$l = 296\, metres$$

Low Resistivity Materials And Their Applications:-

Copper

Properties :

1. Pure copper is one of the best conductors of electricity and its conductivity is highly sensitive to impurities.
2. It is reddish-brown in colour.
3. It is malleable and ductile.
4. It can be welded at red heat.
5. It is highly resistant to corrosion.
6. Melting point is 1084^0C.
7. Specific gravity of copper is 8.9.
8. Electrical resistivity is 1.682 micro ohm cm.
9. Its tensile strength varies from 3 to 4.7 tonnes/cm^2.
10. It forms important alloys like bronze and gun-metal.

Uses : Wires, cables, windings of generators and transformers, overhead conductors, busbar etc.

Hard drawn (cold-drawn) copper conductor is mechanically strong with tensile strength of 40 Kg/mm^2. It is obtained by drawing cold copper bars into conductor length. It is used for overhead line conductors and busbars.

Annealed Copper (Soft Copper) Conductor. It is mechanically weak, tensile strength 20 Kg/mm^2, easily shaped into any form.

Low-resistivity Hard Copper. It is used in power cables, windings and coils as an insulated conductor. It has high flexibility and high conductivity.

➢ **Silver**

Pure silver has high electrical conductivity and corrosion resistance. It is used where high resistance is not required. In order to make it hard 15% of copper is added into it. To make it more hard for use in commutator segment of DC motors as alloy of a silver-copper containing 40% of copper is used. For brushes and collector ring motors silver graphite alloy containing a small percentage of graphite is used because it provides sliding lubrication.It is best known electrical conductor.

Properties

1. It is very costly.
2. It is not affected by weather changes.
3. It is highly ductile and malleable.
4. Its resistivity is 165 micro ohm cm.

Uses : Used in special contact, high rupturing capacity fuses, radio frequency conducting bodies, leads in valves and instruments.

➢ **Aluminium** -Aluminium is widely available in India and is used extensively in the field of electrical engineering. So far as electrical conductivity is connected, it is next best to copper. Its resistivity is 2.8×10^{-8}ohm-m. It can be drawn into the wires. Aluminium is soft metal but when alloyed with some other material like magnesium, Silicon or iron it requires higher mechanical strength and can be used for overhead transmission lines.

Aluminium is quite extensively used for flexible wires overhead transmission lines, bus-bars, squirrel cage induction motor rotor bars and in many other applications

Properties:

1. Pure aluminium has silvery colour and lustre. It offers high resistance to corrosion. Its electrical conductivity is next to that of copper.
2. It is ductile and malleable.
3. Its electrical resistivity is 2.669 micro ohms cm at 20^0C.
4. It is good conductor of heat and electricity.
5. Its specific gravity is 2.7.
6. Its melting point is 658^0C.
7. It forms useful alloys with iron, copper, zinc and other metals.
8. It cannot be soldered or welded easily.

Uses : Overhead transmission line conductor, busbars, ACSR conductors. Well suited for cold climate.

➢ **Steel**.

Steel contains iron with a small percentage of carbon added to it. Iron itself is not strong but when carbon is added to it, it assumes very good mechanical properties. The tensile strength of steel is higher than that of iron. The resistivity of steel is 8-9 times higher than that of copper. Hence, steel is not generally used as conductor material. Galvanised steel wires are used as overhead telephone wires and as earth wires. Aluminium conductors are steel-reinforced to increase their tensile strength.

➢ **<u>Gold :-</u>** Gold is the best known electrical conduction. It is found of all-over the world but not a sufficient to make it economical. Gold is generally found veins among rocks and ores of other material. It is also found in the form of dust in the beds of rivers gold has a density of 19.3 times that of water at 20°C it melts at 1063°C and boils at 2100°C. It is malleable and ductile and can be easily beaten into translucent sheets as thin as 0.00001 mm. It is largely used as alloy to make coins and jewellery Its good corrosion resistance property makes its alloy very much useful as corrosive resistant brazing material.

Example- Calculate the diameter of copper wire of length for meters used as winding material in a temperature such that the resistance of the whole winding is 2 ohms. Calculate the diameter of the wire if aluminium is to be used for the above winding, resistance remaining the same.

Solution- We know:

$$R = \rho \frac{l}{a}$$

Putting the values for copper wire we get :

$$2 = \frac{1.7 \times 10^{-8} \times 100}{a}$$

$$\therefore \quad \alpha = \frac{1.7 \times 10-8 \times 100}{a} = 0.85 \text{ sq-mm.}$$

Area, $a = \frac{\pi}{4} \times d^2$, where d is the diameter of the wire

$$\therefore \quad d = \sqrt{\frac{a \times 4}{\pi}} = \sqrt{\frac{0.85 \times 4}{3.14}} = 1.05 \text{ mm.}$$

For aluminium wire :

$$2 = \frac{2.8 \times 10^{-8} \times 100}{a}$$

$$a = \frac{2.8 \times 10^{-8} \times 100}{2} = 1.4 \text{ sq. mm.}$$

$$d = \sqrt{\frac{1.4 \times 4}{3.14}} = 1.34 \text{ mm.}$$

So the diameter of the aluminium wire is $\frac{1.34}{1.05}$

$= 1.28$ times that of copper wire.

Therefore, the space occupied by the aluminium winding will be more and hence the size of the transformer will be bigger.

<u>Standard Conductors</u>:- Standard conductor is made by twisting the wire (stands) together to form layers. Generally, Stranding is done in opposite directions for successive layers. This means if the wire of one layer is twisted in left-hand direction the next layer of wire will be twisted in the right hand direction and so on.

A standard stranding consists of 6 wines around 1 wire then 12 wires around the previous 6, then 18 wires around the 12, then 24 wires around the 18 and so on. The number of layers to be provided will depend on the number of wires to be provides Note that the central wire is not counted as a layer. Instead of a single wire, three or four standard wires may also be put in the centre and over them layers may be formed. It 3 standard wires are put in the Centre, 9 wires will be in the first layer than 15 wires in the second layer and so on. In each of the above case, the increase in the number of wires in each successive layer is 6.

Number of wire in centre	1 Wire	3 Wire	4 Wire
Number of wires in nth layer from centre	6n	3+6n	4+6n
Total number of wires in a standard conductor having n layer	$1 + 3n(1 + n)$	$3(1 + n)^2$	$(4 + 3n)(1 + n)$
Diameter over the nth layer in centimetre where d increases the	$(1 + 2n)d$	$(2.155 + 2n)d$	$(2.414 + 2n)d$

diameter of ach wire in centimetres			

Bundle Conductors:- "The adoption of bundled conductors in extra high tension power transmission enables standard conduction to be employed and gives an increased current carrying capacity compared with a single conductor of equivalent the voltage stress at the Conductor surface is reduced by using bundled Conductor corona loss is smaller and the time is less enable to Cause Radio interference.

Low Resistivity copper alloys :- We have noticed earlier that copper becomes mechanically hand when it is drawn. however, hardening of copper can also be done by alloying with other materials.

Brass :- When copper alloyed with tin (60%. Coppers & 40% of zinc) it is called Brass, Brass has high tensile strength but has lower conductivity than copper. It can be easily shaped by pressing and it lends itself to deep drawing.

Bronze :- Copper when alloyed with tin (8% to 16%) and a very small percentage of a third element like Cadmium, Beryllium, Phosphorus, Silicon, etc. is called Bronze. Bronzes are given their name based on the element which is added to copper and tin to form the alloy for example, when the third element is phosphorus, the alloy is called phosphor bronze. If the third element is silicon or cadmium, the alloy is called Silicon bronze or cadmium bronze, respectively. All bronze possess high mechanical strength as compared to copper but have lower conductivity. This is more corrosion resistance than zinc, So bronzes are more free from corrosion than brasses. Cadmium Bronze is use for contacting conductor and commentator segments. Beryllium bronze whose mechanical strength is higher than cadmium bronze which is used for making current carrying springs, sliding contacts and knife switch blades, etc.

Beryllium Copper alloy:- The copper alloy containing beryllium is also called Bronze. It has high conductivity and mechanical strength. Its hardening and elasticity property can be changed by giving appropriate heat treatment. It is used for making current carrying spring, brush holders, bellows, coil spring, sliding contacts and knife switch blades.

High Resistivity springs Materials and Their Applications :-

Common example of high resistivity materials are Manganine, Constantan, nichrome, Tungsten etc. Depending upon the application one of the specific above mentioned high resistivity materials is chosen to meet the specific requirements of the given application. For example- materials used in making wire wound precision resistance and shunts for measuring instruments, resistance boxes, coils for precision electrical measuring instruments and the like should have negligible temperature Coefficient of resistance and should be draw able.

Tungsten

Properties :

1. It is grayish in colour when in metallic form.
2. It has a very high melting point (3300^{0}C)
3. It is a very hard metal and does not become brittle at high temperature.
4. It can be drawn into very thin wires for making filaments.
5. Its resistivity is about twice that of aluminium.
6. In its thinnest form, it has very high tensile strength.
7. It oxidizes very quickly in the presence of oxygen even at a temperature of a few hundred degrees centigrade.
8. In the atmosphere of an inert gas like nitrogen or argon, or in vacuum, it will reliably work up to 2000^{0}C.

Uses : It is used as filaments of electric lamps and as a heater in electron tubes. It is also used in thermionic valves, radars. Grids of electronic valves, sparking and contact points.

➢ **Carbon**.

Carbon is mostly available as graphite which contains about 90% of carbon. Amorphous carbon is found in the form of coal, coke, charcoal, petroleum, etc. Electrical carbon is obtained by grinding the raw carbon materials, mixing with binding agents, moulding and baking it.

Properties :

1. Carbon has very high resistivity (about 4600 micro ohm cm).
2. It has negative temperature coefficient of resistance.
3. It has a pressure-sensitive reistance material and has low surface friction.
4. The current density is 55 to 65 A/cm^2.
5. This oxidizes at about 300^{0}C and is very weak.
6. It has very good abrasive resistance.
7. It withstands arcing and maintains its properties at high temperature.

➢ **Platinum**

Properties :

1. It is a grayish-white metal.
2. It is non-corroding.
3. It is resistant to most chemicals.
4. It can be drawn into thin wires and strips.
5. Its melting point is 1775^{0}C.
6. Its resistivity is 10.5 micro ohm cm.
7. It is not oxidized even at high temperature.

Applications:

1. It is used as heating element in laboratory ovens and furnaces.
2. It is used as electrical contact material and as a material for grids in specialpurpose vacuum tubes.
3. Platinum-rhodium thermocouple is used for measurement of temperatures up to 1600^{0}C.

➢ **Mercury**

Properties:

1. It is good conductor of heat and electricity.
2. It is a heavy silver-white metal.
3. It is the only metal which is liquid at room temperature.
4. Its electrical resistivity is 95.8 micro hom cm.
5. Oxidation takes place if heated beyond 300_0C in contact with air or oxygen.
6. It expands and contracts in regular degrees when temperature changes.

Uses : Mercury vapour lamps, mercury arc rectifiers, gas filled tubes; for making and breaking contacts; used in valves, tubes, liquid switch.

Super Conductivity:- There are some metals and mechanical compounds whose resistivity becomes zero when their temperature is brought near 0° Kelvin (-273°C). At this stage such metals or compounds are said to have attained superconductivity. For example - Mercury becomes super conducting at approximately 4.5° Kelvin (268.5°C). Super conductivity was discovered by Heike Kamerlingh Onnes at the University of Leiden in the Netherlands in 1911.

There are two types of superconductors commonly known type I and Type II superconductors. Type I Superconductors are they have soft superconductors. They are usually pure specimens of some elements i.e, metals they have very little use in technical applications. Whereas Type II super conductors are hard super conductors, they are usually alloys of metals with high value of resistivity in normal state. These are very useful as compared to type I super conductors.

Superconducting Materials:- Many metals and compounds have Superconducting property at very low temperatures. Super conductivity has been observed to occur in poorer conductors such as tin, lead and tantalum rather than in better conductors such as gold silver and copper.

It has been found that super conductors may not only be pure metal but various alloys and chemical compounds as well. At present about 30 superconductor metals and more than 6oo superconductor alloys are already known. The highest temperature at which until now, super conductivity has been observed to occurred is 20°k (-253°c) for a compound consisting of Niobium, Aluminium and Germanium.

Application of Superconductor Materials-

Electrical Machines:- Efforts are being made at present to develop electrical machines and transformers utilizing super conductivity. Calculations show that if we could use super conductors as conducting materials in addition to super conducting magnets which are already being produced. It is possible to manufacture electrical generators and transformers in exceptionally small size having an efficiency as high as 99.99%

Power Cables:- Super conducting material it used for power cables will enable transmission of power over very long distances using a diameter of a few centimeters without any significant power loss or drop in voltage.

Electromagnets:- Superconducting solenoids which do not produce any heat during operations have been produced. However it must be noted that superconductivity can be destroyed it the magnetic field exceeds a critical value. It has been possible to design electromagnet using superconductivity for use in laboratories and for low temperature devices like the maser.

Future Prospects:- It must be realized that the above applications require the conductor to be maintained at temperature very close to 0°k. This may often mean that the whole equipment associated with the conductor has to be kept at near 0ºk. This is a great challenge facing the scientists today. Indeed a new technology known as Cryogenics has been developed to tackle this problem

Presently Helium is used to achieve low temperature required for super conductivity. Helium being expensive gas efforts are being made to develop compounds which exhibit superconductivity at temperatures possible to be obtained by the more easily available and cheaper than Hydrogen gas.

Semiconducting materials

Introduction

"A semiconductor material is one whose conductivity lies between that of a conductor and an insulator." The two most commonly used semiconductor materials are germanium and silicon.

1.3 Conductor, Insulators and Semiconductors

Any solid is formed by bonding between atoms. Inter-atomic bonds are of three main types:

The first bond is the ***metallic bond***. In this type, the atoms of the elements which have 1, 2 or 3 valence electrons, being loosely held, give up those electron to form an electron cloud in the space of the atoms and become positive ions. The material is held together by electrostatic force between positive ions and electron cloud. The elements having small number of valence electrons are formed by this type of bonding and become ductile and have good conduction of electricity. These elements are known as ***conductors***.

The second one is the ***ionic bond*** where the atoms of different elements transfer electrons from one to the other so that both have stable outermost orbits. For example, in sodium chloride, sodium atom gives out its one valence electron to chlorine atom and both become stable with 8 electrons in outermost orbits. At the same time, one becomes positive ion and the other negative ion. The electrostatic force between the two gives rise to the bonding. High hardness and low conductivity are typical properties of ionic bond. Therefore these materials are ***insulators***.

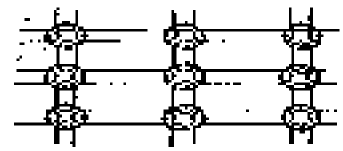


Fig 3.1

The third bond is called ***covalent bond***. In this bond, the atoms of the materials having 4 or more valence electrons share their electrons with neighbouring atoms as shown in Fig.3.1 The atoms of such materials behave as if they have full outer orbits.

This gives full strength to the material and low electrical conductivity because no electrons are free to move. Certain materials allow valence electrons to become free by thermal energy. These elements are known as ***semiconductors***.

An atom is identified by its atomic number which indicates the number of protons in the nucleus (or the number of electrons in the orbits). For example, an oxygen atom has 8 protons and 8 neutrons in the nucleus and 8 orbital electrons. Therefore, its atomic weight is 16 and atomic number is 8.

1.5 Electron Energy and Energy Band Theory

When each atom with its neighbouring atom shares electrons in order to fill its valence ring with 8 electrons, a covalent bond is formed. Figure-3.2 shows covalent bonding.

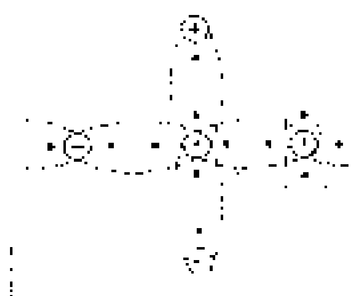


Figure -3.2 Set of covalent bond

When atoms enter into this bonding, each atom in effect has 8 valence electrons and this results in making such material a good insulator. Covalent bonding leads to the development of a poly crystal. In a poly crystal, several individual crystals are held

together imperfectly. The extra atoms are not properly locked in place. Due to impurities, there may be extra electrons which cannot lock into the covalent bond structure. Thus, a semiconductor is produced.

An impure material having three valence electrons is called trivalent bond, e.g. Gallium, Indium and Aluminium.

An impure material having five valence electrons is called trivalent bond,

e.g. Antimony, Arsenic, Phosphorous.

Excitation of Atoms

When each electron in an atom is in its normal orbit, the atom is said to be in an unexcited state.

To move an electron further away from its nucleus requires additional energy. The additional energy can be obtained from any of the following sources: light, heat static electricity, magnetism, kinetic sources.

When the electron is in the higher energy level, the atom is said to be in an excited state. The quantum of energy, in electron volts, required to move an electron from one energy level to higher energy level varies from material to material.

When the required amount of light or heat energy is absorbed by a valence electron, it will leave the valence bond and move up to the ionization level. If it does so, it is released from the attraction forces of the nucleus. Then it is free to float between the atoms and to conduct electricity. An electron above ionization level is said to be in the conduction band and is called a free electron.

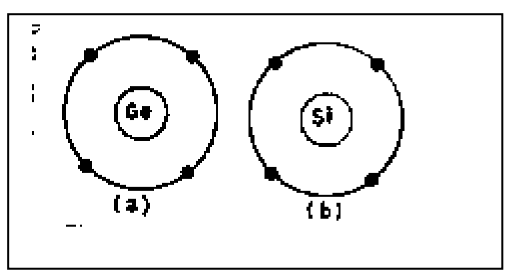
When the electron leaves the valency bond, the resulting atom is no longer neutral but has a positive charge and is called positive ion. The atom is said to be ionized.

The atom that has been ionized by the loss of an electron, does not remain so for a long time. Its positive charge will attract a nearby free electron which will give up its acquired energy. Thus, there is a constant interchange of electrons being given up and retrieved.

Energy Band Representation of lonization

In the silicon atom, K and L shells are full, but M shell contains only four electronics. According to the $2n^2$ formula, the M shell can contain 18 electronics, but the M shell in silicon is the valence shell and thus can have not more than 8 electronics.

Figure -3.3 (a) simplified silicon and germanium atom



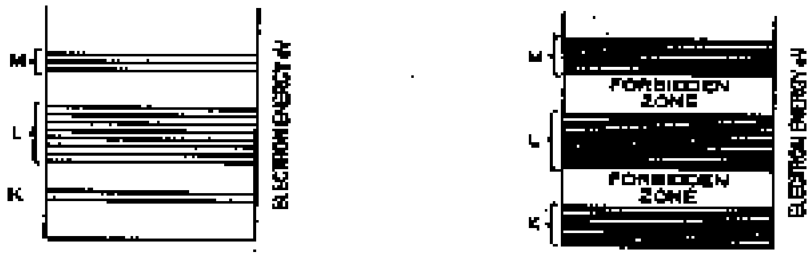


Figure -3.3 (b) Energy band representation of ionisation

In the germanium atom, the K, L and M shells are filled and the N shell is the valence shell containing 4electronics. Since only the valance electrons are important from the chemical and electrical point of view, both germanium and silicon atoms are shown in simplified from by representing only the outer most shell in Figs. 3.3 (a) and (b).

Simplified Si and Ge Atoms

The electrical characteristics of a semiconductor fall between those of a conductor and an insulator.

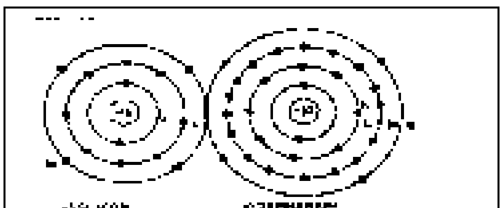


Fig- 3.4 Shell arrangement

A semiconductor has 4 electrons in its valence ring (outmost orbit). A good insulator has 8 electrons in its valence ring. The best conductor has one electron in the valence ring.

The two most widely used semiconductors are silicon (si) and germanium (Gi). Their atoms structure are shown in Figs. 3.4 (a) and (b).

N-type Material.

When a pentavalent impurity is added to an intrinsic material such as silicon or germanium, only four of its valence electrons lock into the covalent bond formation of atom structure. The fifth valence electron of the impurity atom is free to wander through the crystal.

Fig- 3.5 Arsenic impurity atom provides a fifth electron

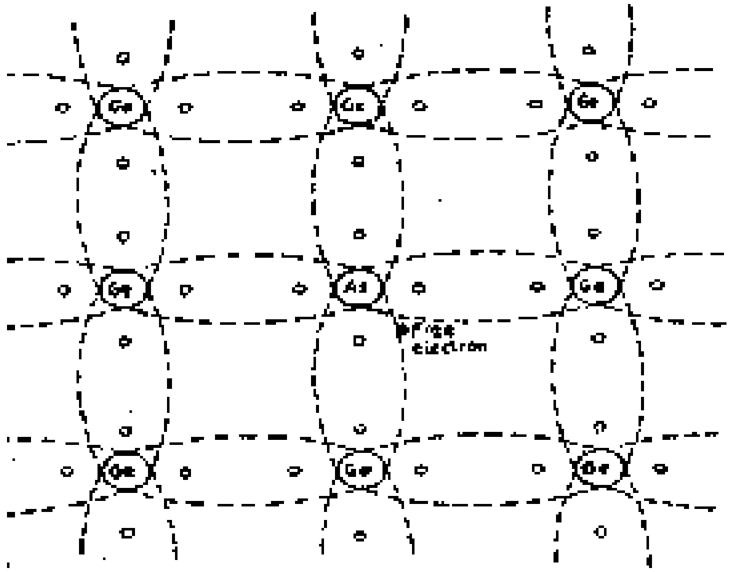


Figure 3.5 shows the addition of an atom of arsenic as an impurity. The impurity atom becomes ionized and has a positive charge when its fifth electron moves away. The positive impurity ion is not free but is firmly held in the crystal structure. The pentavalent atom donates an extra electron and is called a donor impurity. A material doped with a donor impurity has excess of electrons in its structure. It is called N-type material. The net charge of N-type material is still natural since the total number of electrons is equal to the total number of protons.

Arsenics impurity atom provides a fifth electron that cannot enter a covalent bond structure.

P-type Materials.

When a trivalent impurity is added to the intrinsic material, the two lock into a crystal structure. The impurity has three valence electrons. There is a hole in the covalent bond structure created by the lack of an electron. The hole represents an incomplete covalent bond and exhibits a positive charge. In order to complete the bond and from a stable 8-electron structure, a valence electron from a nearby atom gains sufficient energy to break loose from its bond and jumps into the hole due to its attraction. Therefore, this type of impurity is called an “acceptor”. The electrons available to fill the hole and complete the bond have been release by the nearby atom whose bonds have been broken and hole created. Thus, the process will continue creating a mobility of holes. The impurity atom becomes negatively ionized as accepts an electron. The germanium or silicon atom which releases one electron become positively ionized. The net charge of the material is still neutral. The total number of electrons is equal to the total number of protons.

Semiconductors Commonly Used

The following materials are commonly used as semiconductors:

(i) Boron
(ii) Carbon
(iii) Silicon
(iv) Germanium
(v) Phosphorus
(vi) Arsenic
(vii) Antinomy
(viii) Sulphur
(ix) Selenium
(x) Tellurium
(xi) Iodine

Intrinsic Semiconductors.

If a crystal (silicon or germanium) does not contain any impure atoms (contains only one type of atoms), it is called an intrinsic material. When an electron is freed from the atom of an intrinsic material, it breaks a covalent bond and leaves behind a vacancy (called a *hole*). The free electron and the hole form an electron-hole pair. The higher the temperature, the greater the number of free electrons and holds. When a voltage is applied to an intrinsic material, it acts as a conductor.

Extrinsic Semiconductors.

Pure silicon or germanium exhibits characteristics closer to that of an insulator than a semiconductor. In order to make a material conducting, a small quantity of impurity must be added to it. The addition of impurity makes pure germanium or silicon a conductor. The process of adding impurities is called “doping”.

The extent to which the impurity has been added is called the “doping level”. When a pentavalent group provides an extra electron to the semiconductor material, the atom of the material which donates the extra electron is called a “donor atom”

When a trivalent group is added to intrinsic materials such as silicon, one covalent bond is broken, that is, a hold is created. An electron from an adjacent atom can fill the hole which is now moved to another atom. The doping atom has now one surplus negative charge and has become a negative ion. A hole is the absence of an electron and hence has a positive charge. The doping element is an “accepted”, since it takes or accepts an electron.

Majority and Minority Carriers.

In N-type material, conduction takes place through the electrons created mostly by the doping and a small number created by thermal generation.

The small number of holes created by thermal generation ove in opposite direction. In N-type material, the number of free electrons is large. These electrons are called majority carriers. Holes are in small numbers and are called minority carriers.

In p-type material, the holes are majority carriers and electrons are minority carriers.

➢ Working and Application of Semiconductors

Semiconductor materials are used in :

(i) Rectifiers

(ii) Temperature-sensitive resistors

(iii) Photoconductive and photovoltaic cells

(iv) Varistors

(v) Hall effect generators

(vi) Strain gauges

(vii) Transistors

(viii) LDR and LCD

Some of them are discussed below

Germanium and Silicon Rectifiers. When a P-type material and an N-type material are joined together, they form a junction called P-N junction.

When an external voltage is applied across the two material, a flow of current results if the positive and negative terminals of the voltage source are connected respectively to the ends of the P and N material. The voltage applied this way is called “forward-biasing” the P-N junction. If the applied voltage is reversed, that is, the positive of the supply voltage is connected to N side and negative of the supply is connected to the P side, there is no flow of current. This is called “reverse biasing”. Thus the P-N junction offers high conductivity when forward biased and no conductivity when reverse biased. Thus, the semiconductor can be used as a rectifier. The modern P-N junction rectifiers use germanium or silicon material. Circuit diagram Fig. 3.6 a & b - below also illustrate the characteristics.

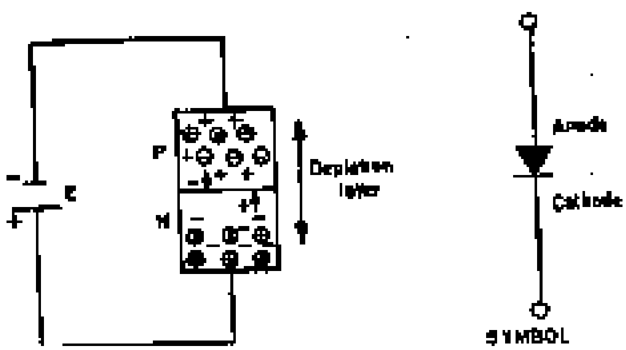
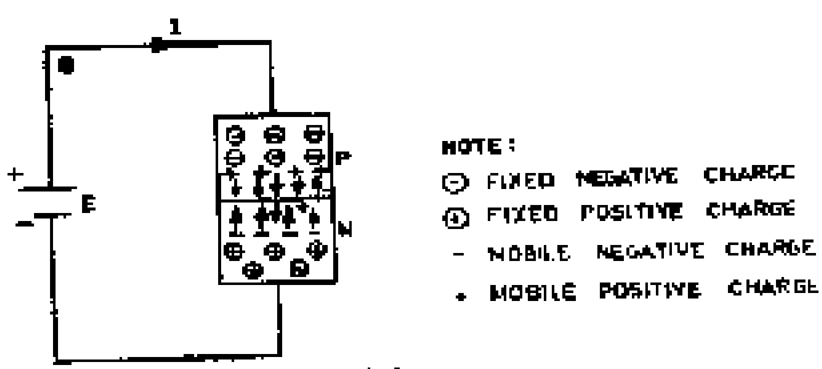


Fig-3.6 a The operation of PN junction under reverse bias



(b)

Fig- 3.6 b The operation of PN junction under forward bias

- **<u>Copper Oxide Rectifiers</u>**. The earliest semiconductor to be used was copper oxide. Its application was in copper oxide rectifier.

Copper oxide rectifier is a place of 99.98 % pure copper on which a film of fuprous oxide is produced by a special process. From one side of the plate, cuprous oxide is cleaned and electrode is soldered directly to the copper. The second electrode is soldered to cuprous oxide film. When a positive potential is applied to the oxide layer and negative to the copper, it corresponds to forward biasing of a P-N junction. By arranging the copper plate elements in stacks, rectifiers for use in many kinds of measuring instruments and circuits can be obtained. These rectifiers have low permissible current density. They are not used for power supply purposes.

To have a good contact with copper oxide, a lead plate is pressed against it. The two terminals of the rectifiers are the copper plate and lead plate. The oxide will be in between the plates as shown in figure- 3.7. This rectifier will allow the current to flow only from oxide to copper and will not allow flow from copper to oxide.

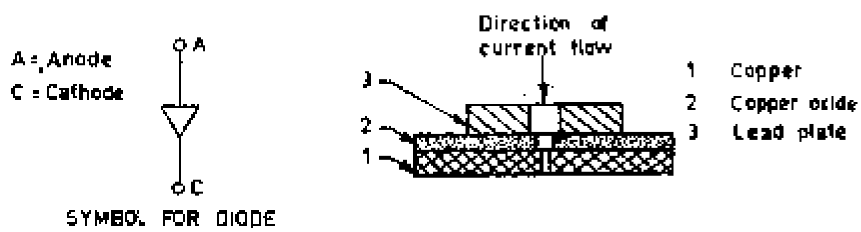


Fig- 3.7 Copper oxide rectifier

The voltage that may be applied to a single rectifier ranges between 4 and 8 V, so a number of units are connected in series for operating on high voltages. Similarly, parallel connected of the units, increases the current rating of the rectifiers, as the maximum current density in the forward direction is 0.1 to 0.15 A/cm^2 at an allowable voltage of 8 V.

The life of copper oxide rectifiers is 12 to 15 years and efficiency is 70 %.

Applications : These types of rectifiers are mostly used for meters, battery cell charging, X-ray works, measuring instruments, railway signaling, telecommunication systems, etc.

➢ **Selenium Rectifiers**.

In this type, a film of 0.5 mm. thickness is deposited on one side of the metallic back plate (iron or aluminium). By means of chemical treatment, a film of “blocking” or “barrier” layer is formed between selenium and counter electrodes.

The rectification is from back plate to selenium. The rectifier construction is as shown in figure-3.8.

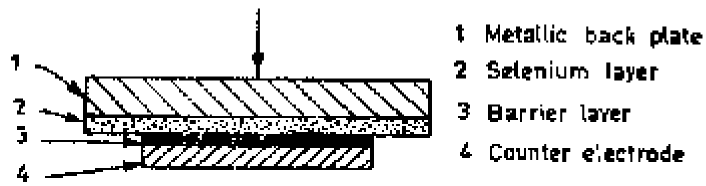


Fig-3.8 Selenium rectifier

A single unit can sustain 6 V. The normal current density is about 0.04 A per cm^2 for full wave rectification. The power efficiency is 50 to 75 %.

The units can be combined in series or in parallel, similar to that of copper oxide rectifiers to work at desired voltage or for the required current capacity.

Applications : This type of rectifiers are widely used for battery charging, telegraph and telephone circuits, control circuits, railway signaling, meters, electroplating and other works.

Such rectifiers are available in capacities of up to 50 to 100 KW.

➢ **Temperature-sensitive Elements (Thermistors)**

If the temperature of a semiconductor material is increased, that causes a decrease in its resistance. This property is used in temperature sensitive elements which are called as „thermistors‟.

The termistors are thermally sensitive material (resistors). They are made from oxides of certain metals such as copper, manganese, cobalt, iron and zinc.

Applications of thermistors: Thermistors find application in temperature measurements and control. They sense temperature variations and convert these variations into an electrical signal which is then used to control heating devices.

Thermistors are also used for measurement of radio frequency power, voltage regulation and time delay circuits.

- **Photoconductive Cells**

The resistance of semiconductor materials is low under light and increases in darkness. Phtoconductive cells can be used in applications which require the control of a certain function or event according to the colour or intensity of light.

Applications: They are used in burglar alarms, flame detectors and control for street lights.

- **Photovoltaic Cells**

Photovoltaic cells are devices that develop and emf when illuminated. They convert light energy directly into electrical energy.

Applications: The applications of photovoltaic cells are in photographic exposure meters, lighting control systems, automatic aperture control in cameral.

- **Varistors**

The resistance of semiconductors varies with the applied voltage. This property is used in devices called varistors.

Applications. They are used in voltage stabilizers and for motor speed control.

- **Hall Effect Generators**

When a current flows through a semiconductor bar placed in a magnetic field, a voltage is developed at right angles to both current and the magnetic field. This voltage is proportional to the current and the intensity of the magnetic field. This is called the "Hall effect".

Consider the semiconductor bar shown in Fig.3.9, which has contacts on all four sides. If a voltage E_1 is applied across the two opposite sides A and B_2 a current will flow.

If the bar is placed perpendicular to magnetic field B as shown in the figure, an electrical potential E_H is generated between the other two contacts C and D. This voltage E_H is a direct measure of the magnetic field strength and can be detected with a simple voltameter.

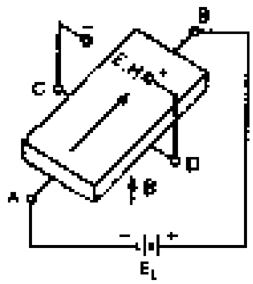


Fig- 3.9

Hall Effect Device

Applications. The hall effect generators may be used to measure magnet is fields. It is capable of measuring magnetic field strengths that have a strength of 10^{-6} of the magnetic field of the earth.

- **<u>Strain Gauges</u>**

Semiconductors are sensitive to heat, voltage and magnetic field; they are also sensitive to mechanical forces. If a long thin rod of silicon is pulled from end to end, its resistance increases considerably because the mechanical force pulls each silicon atom slightly away from its adjacent atom. This increases the breadth of the forbidden energy gap, which increases the resistivity of the rod. Silicon and other semiconductors are used in strain gauges.

Applications: Strain gauges are used to find the small changes in length of solid substances or objects.

Insulating Materials

Introduction:

For safe and satisfactory operation of all electrical and electronics equipment insulator plays important role. Basically current carrying wires, surfaces need to be covered with insulating material. Let us see the structure of the material on the basis of energy band. In this type of material, the highest occupied energy band (Valence Band) is completely filled. The next higher band (Conduction Band) is quite empty.(Fig.1) The gap between these two bands is too large. When the electric field is applied across these materials, the electrons from valence band cannot reach the conduction band and conduction of electron stops. Such materials are known as insulators. Diamond is an example of this kind of material with a separation of nearly 6eV between valence band and conduction band.

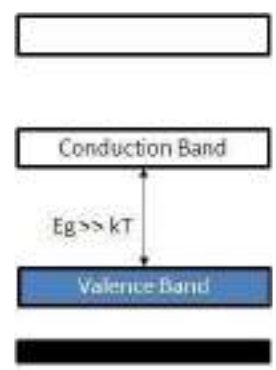


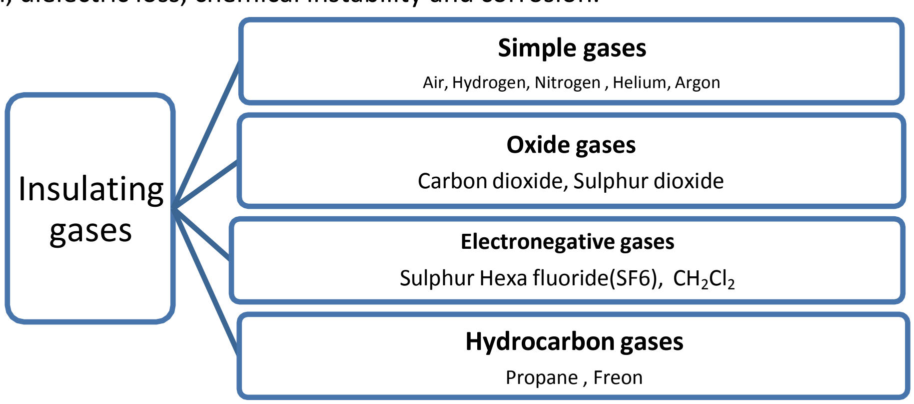
Figure 1. Energy band of INSULATOR

Insulating Materials for Electrical Engineering

The insulating materials used for various applications in electrical engineering are classified in three categories:

- Insulating gases
- Liquid insulating material
- Solid insulating material

1. ***Insulating Gases***: Many gases are used as the medium of heat transfer. All known gases are dielectric in pure form, but from electrical engineering point of view these are classified on the basis of different properties like dielectric strength, dielectric loss, chemical instability and corrosion.



2. **Liquid Insulating Material**: These materials are used for dielectric purpose to eliminate the air and other gases Insulating liquids are organic liquids used as coolant. These are categorised according to temperature range where they are used. It is used in transformers, circuit breaker, bushings, cables, capacitors etc. along with solid insulants to operate with an acceptable performance. An ideal insulating liquid material must have following properties:
    - High dielectric strength, impulse strength and volume resistivity.
    - Low dielectric dissipation factor.
    - High or low dielectric constant.(depending upon application)
    - High specific heat and thermal conductivity.
    - Excellent chemical stability and gas absorbing properties.
    - Low viscosity, density, volatility and solvent power and high flash point.
    - Good arc quenching properties.
    - Non-flammable and non toxic

Table 1 *Liquid Insulating Materials*

Type of Liquid	Temperature Range	Applications
Petroleum oils (Mineral oils)	*-50 to 110^{0}C*	All types
Askarels	*-50 to 110^{0}C*	Transformer, Capacitor, Switch gear
Silicon Liquids	*-90 to 220^{0}C*	Transformer,
Halogenated Hydro carbon	*-50 to 200^{0}C*	Electric equipment
Synthetic hydro carbon	*-50 to 110^{0}C*	Cables and Capacitors
Organic esters	*-50 to 110^{0}C*	Electronics equipments
Vegetable oil	*`-20 to 100^{0}C*	Limited application

General Properties of Insulating Material :

The suitability of an insulating material for a specific purpose use can be decided by knowing its different properties. So we have to know the exact requirement of the application and the required property hold by the insulating material. Based on uses in different applications following properties of materials are useful.

1. **Electrical Properties:** The insulating material used in electrical or electronics appliances, should be considered for following:
    - Insulation resistance
    - Dielectric constant or permittivity
    - Breakdown voltage or dielectric strength
    - Dielectric loss

1. 1 Insulation Resistance:

This is the ohmic resistance offered by an insulation coating, cover or material in an electric circuit which tends to produce a leakage current through the same with an impressed voltage across it.

Let us consider a cable of inner and outer radii r_1 and r_2, length l and resistivity of insulating material ρ. Considering a very thin layer of radial thickness *dr* at a radius *r*, the length through which the leakage current flows is *dr* and area of cross section provided to flow of current is *2πrl*.

Hence insulation resistance of the layer under consideration $= \frac{\rho dr}{2\pi r l}$

Insulation resistance of the cable can be determined by integrating above expression between the limits r_1 and r_2. Insulation resistance of the cable is given by,

$$R = \int_{r1}^{r2} \frac{\rho dr}{2\pi r l} = \frac{\rho}{2\pi l}\int_{r1}^{r2} \frac{dr}{r} = \frac{\rho}{2\pi l} log_e \frac{r_2}{r_1}$$

The equation states that, the resistance of the cable decreases with increase in length.

1. 2 Dielectric constant or Permittivity:

The permittivity of the insulating material varies with temperature and frequency in some cases. The materials like HCl, H_2O, CO, NH_3 have permittivity variation with change in temperature.

1. 3 Dielectric strength:

It is the maximum impressed voltage bearing capacity of insulator per unit thickness of material, up to which current does not flow through it. When current flows through the insulator is known as dielectric failure.

The dielectric strength of an insulating material decreases with the duration of time the voltage is applied, moisture, contamination, high temperature, heat ageing, mechanical stress etc. and decreases up to 10% of laboratory values.

1. 4 Dielectric loss:

Dielectric losses occur in all solid and liquid dielectric due to: a conduction current and hysteresis.

- The conduction current is due to imperfect insulating qualities of the dielectric and is calculated by the application of Ohm's law. It is in phase with the voltage and results in the power loss (I^2R) in the material, which is dissipated as heat.

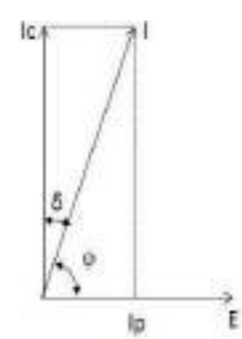


Figure 2. Plot of E against Ic

- Dielectric hysteresis is defined as the lagging of electric flux behind the electric force producing it so that under varying electric forces a dissipation of energy occurs. The energy loss due to above cause is called the dielectric hysteresis loss. The energy is dissipated as heat. This loss gives an indication of the amount of energy absorbed by the material, when subjected to AC fields.

2. Visual Properties:

An insulating material possessing two opposite properties: transparency and thermal insulation is suitable in case of reduction of energy consumption for heating and air conditioning and electrical energy savings. This is known as visual properties. Study of appearance, color and crystalline structure are the measures of this property. Glass, Aerogel hold the required visual properties. **Aerogel** is used in case of highly energy efficient windows.

3. Mechanical Properties:

Mechanical properties such as tensile strength, impact strength, toughness, hardness, elongation, flexibility, mechanical strength, abrasion resistance etc. are to be considered for choosing the insulating material.

3. 1. Mechanical Strength:

The insulating material should possess sufficient mechanical strength to respond mechanical stress. Mechanical strength is affected by following factors.

- Temperature rise: It badly affects the mechanical strength of the insulating material.
- Humidity: It is the climatic effect which affects also the mechanical strength.
- Porosity: An insulating material of high porosity will absorb more moisture and thereby affects the electrical properties as well as mechanical strength.

4. Thermal Properties:

Following thermal properties are considered for selecting insulating material of different applications.

4. 1. *Thermal stability*: The insulating material must be stable (no change in physical state) within the allowed temperatures. Certain materials like wax and plastic get soft at moderate temperatures. So the mechanical property of the material is affected. Hence the operating temperature of the material is to be noted before its use.

4. 2. *Melting point*: The insulating material should have melting point (temperature bearing capacity without being melt), above that of operating temperature.

4. 3. *Flash point*: This is an important property of insulating oils used in transformer. Flash point of a liquid insulator is that temperature at which the liquid begins to ignite.

4. 4. *Thermal conductivity*: In electrical appliances heat is generated during operation, which should be transferred to atmosphere, to maintain the operating temperature within the limit. Hence the insulators should have very low thermal conductivity

4. 5. *Thermal expansions*: Rapid and repeated load cycle on electrical appliances cause corresponding expansion and contraction of the insulators. In a result voids are created and affect the breakdown phenomenon. Thus two insulating material of different coefficient of thermal expansion should be wisely selected.

4. 6. *Heat Resistance*: The insulating material used must be able to withstand the heat produced due to continuous operation and remain stable during the operation. At the same time it should not damage the other desired properties.

5. Chemical Properties:

Certain chemical properties are also required to be considered for the insulating materials.

5. 1. *Chemical Resistance*: It is the ability of the insulating material to fight against corrosion in the presence of gases, water, acids and alkalis. For materials which are subjected to high voltage, high chemical resistance is also necessary.

5. 2. *Hygroscopity*: Many insulating materials are hygroscopic. Sometimes the insulation may come in direct contact with water. The porous materials are more hygroscopic than dense ones. Small amount of moisture absorbed by an insulating material affects its electrical properties drastically.

5. 3. *Moisture Permeability*: The tendency of an insulating material to pass moisture through them is known as moisture permeability. Moisture can penetrate through very small pores as the size of water molecule is very small. So this property is vital for selecting the protective coating, cable sheaths etc.

6. **Ageing:** Ageing is the long term effect of heat, chemical action and voltage application. These factors decide the natural life of insulators and hence of an electrical apparatus.

Insulating Gas: Properties and applications

Air: Air provides insulation between the over-head transmission lines. It is the best insulating material when voltages are not very high. It is also used in air capacitor, switches and various electrical equipments.

It is easily available, non-inflammable, non-explosive, small dielectric strength (nearly 3to 5 kV/m) and reliable at low voltage.

Hydrogen: It is commonly used for cooling purpose in electrical machine due to its lightness. Its high thermal conductivity helps to transmit heat from windings of high capacity alternator. Thus it reduces windage losses and increases efficiency.

Nitrogen: Nitrogen is used in place of air, to prevent oxidation due to its chemically inert property. It is generally used in transformers, gas pressure cable and capacitors.

Carbon Dioxide: Carbon dioxide is used in certain fixed type capacitor, and is used as a pre-impregnate for oil filled high voltage apparatus, such as cables and transformers. The relative permittivity of carbon oxide is 1.000985 at 0^0 C.

Sulphur Hexafluoride (SF_6): The electromagnetic gases have high dielectric strength compared to other traditional dielectric gases like nitrogen and air. The dielectric strength of SF_6 is 2.35 times more than air. The electronegative gases are non-inflammable and non-explosive. The most important gas under this category is sulphur Hexafluoride, while others are Freon gases.

SF_6 is mostly used in high voltage application and its use is most satisfactory in dielectric machines, like X-ray apparatus, Van de Graff generators, voltage stabilizers, high-voltage switch gears, gas lasers etc. SF_6 bears some special properties as follows:

- SF_6 is colourless, nontoxic and non-inflammable gas. It is the heaviest gas and has low solubility in water. The gas can be liquefied by compression. Its cooling characteristic is better than air and nitrogen.
- Under normal temperature conditions it is chemically inert and completely stable with high dielectric strength.
- This gas has very good electronegative property. Its relatively large molecules have a great affinity for free electrons, with which they combine making the gas-filled break much more resistant to dielectric breakdown.

Liquid Insulating Material: Properties and applications

Mineral oils: The operating temperature range of mineral oil is 50-110^{0}C. These hydrocarbon oils are used as insulating oils in transformers, circuit breakers, switch gears, capacitors etc.

In transformers, light fraction oil, such as transil oil is used to allow convection cooling. Its high flash point is 130^{0}C, so it is able to prevent fire hazard. Highly purified oil have a dielectric strength of 180 kv/mm and if the oil contains polar and ionizing material its dielectric loss increases.The dielectric constant is about 2.3 and therefore it is capable of dissolving only very few substances in it and produce the conducting ions. The TRANSIL oil undergoes oxidation, particularly in in the presence of catalysts such as copper, to form sludge and acids.

Light oils having Saybolt viscosity of 100 seconds at 40^0C, have been used under pressure in oil filled high voltage cables.

More viscous or tacky oils with Saybolt viscosity of 2000 seconds at 40^0C, are generally to impregnate the paper in solid type cable.

Askarels: These are non-inflammable, synthetic insulating liquids, used in temperature range of 50 – 110^0C. Chlorinated hydrocarbons are the most widely used among the askarels because of high dielectric strength, low dielectric constant (4 to 6) and small dielectric loss. They do not decompose under the influence of electric arc and have good thermal, chemical and electrical stability.

Chlorinated hydrocarbons as askarels are used as transformer fluids to reduce fire hazards. Chlorinated diphenyl, penta chloro diphenyl, trichloro diphenyl, hexa chloro diphenyl, trichloro benzene, etc., are the most widely used hydrocarbons or askarels. Askarels are generally used to impregnate a cellulose insulating material, such as paper or press board etc., for its high breakdown strength.

Silicon Fluids: It is used in the temperature range of 90-220^0C and it is clear, water like liquid. It is available in wide range of viscosity and stable in high temperature. They are non-corrosive to metal upto 200^0C and bear excellent dielectric properties in wide range of temperature. So it is used as coolants in radio pulse and aircraft transformers.

Fluorinated Liquids: These are non inflammable, chemically stable oils used in temperature range of 50-200^0C. They provide efficient heat transfer from the winding and magnetic circuits in comparison to hydrocarbon oils and used in small electric and radio devices, transformers etc. In presence of moisture electrical properties are deteriorated.

Synthetic Hydrocarbon oils: Polybutylene, Polypropylene is the example of synthetic hydrocarbon oils. They have similar dielectric strength; thermal stability and susceptibility to oxidation properties are similar as that of mineral oils. The operating temperature range is 50-110^0C. These are used in high pressure gas filled cables and dc voltage capacitors.

Organic Esters: These organic fluids are used in the temperature range of 50-110^{0}C. They have dielectric constant and very low dielectric losses. The dielectric constant ranges from 2 to 3.5. The higher range of 12.8 is obtained in tetra hydro-furyloxalate. These fluids are well suited for use in high frequency capacitors.

Vegetable Oils: These insulating liquids have temperature range of 20-100^{0}C. Drying oils are generally suitable in the formation of insulating varnishes, while non-drying oils are used as plasticisers in insulating resin compositions.

Varnish: It is the liquid form of resinous matter in oil or a volatile liquid. Hence by applying, it dries out by evaporation or chemical action to form hard, lustrous coating, which is resistant to air and water.

It is used to improve the insulation properties, mechanical strength and to reduce degradation caused by oxidation and adverse atmospheric condition.

Classification of insulating materials on the basis of structure

Classification of insulating materials is done on the basis of their physical and chemical structure.

Table 2: Classification of materials on the basis of structure of material:

Classification	Insulating Materials
Fibrous material	Wood, paper, cotton, adhesive tapes
Insulating liquids	Transformer oils, cable oils, silicone fluids
Non-resinous material	Bitumen‟s, wax
Glass and ceramics	Glass, porcelain etc.
Plastics	Molding powder, rubber laminations
Mineral	Mica, micanites
Gaseous	Air, H_2, N_2, Ne, CO_2, SF_6, Hg and Na vapor

Table 3: Property and uses of some common electrical engineering materials

Material	Properties	Uses
Paper and press board	Low dielectric loss, Discharge current is lower	High frequency capacitors
Cotton	Hygroscopic, Low di-electric strength, properties can be improved by impregnation	Winding of small magnetic coil, Armature winding of coil and chokes
Wood	High dielectric constant, Highly hygroscopic, dry wood can bear a voltage gradient of 10kV/inch	Terminal block, wedges of armature winding, operating rods in high voltage switch gears.
Bitumen	Hydrocarbons of jet black colour, highly soluble in mineral oil, Poor insulating property, Low hygroscopic, Acid and alkali resistant	Underground cable
Waxes	Complex organic substance, High insulating property, Low hygroscopic	As impregnated material for paper and cloth insulation, dipping medium coating on conductors.
Glass	Organic material containing oxides, silicate and borate etc. Best insulating property, High resistivity and dielectric strength.	Insulating material to form envelope for internal support in bulbs, valves, X-ray tubes, fuse casings etc.
Ceramics	Hard, Strong, Dense, Unaffected by chemical action, Stable at high temperature, Excellent dielectric properties, weak impact strength	High voltage insulation at elevated temperatures in ovens.
Asbestos	Exhibit fiber structure , Can work at high temperatures, Good tensile strengths	Capacitor dielectric, transistor, hybrid circuit substrates, Electromechanical transducers, Not useful for high voltage, Used as thermal insulators and cables in high temperature.
Rubber	Stretchable, Moisture repellant, Good insulating properties, Good corrosion resistance. Can be obtained as hard rubber, synthetic rubber, butadiene rubber, butyl rubber, chloroprene rubber and silicon rubber.	Used as protective clothing such as boots and gloves, also used as insulation covering for wires and cables. Hard rubber is used in housing for storage batteries, panel board, jacketing material.

Special Solid Insulating materials: Properties and applications

MICA: two kind of mica are used as neutral insulating material in electrical engineering. Those are Muscovite mica and Phlogophite mica.

- ***Muscovite Mica**:* The chemical composition of muscovite mica is $KAI_3SiO_3O_{10}(OH)_2$. It is translucent green, ruby, silver or brown and is strong, tough and flexible. It exhibits good corrosion resistance and is not affected by alkalis. It is used in capacitors and commutators.
- ***Phlogophite Mica:*** The chemical composition of this is, $KMg_3AlSiO_3O_{10}(OH)_2$. It possesses less flexibility. It is amber, yellow, green or grey in colour. It is more stable, but electrical properties are poorer compared to Muscovite Mica. It is used in thermal stability requirements, such as in domestic appliances like iron, hotplates etc.

Polyethylene: It is obtained by polymerization of ethylene. The polymerization is performed in the presence of catalyst at atmospheric temperature and pressure around 100^0C. To obtain heat resistance property polythene is subjected to ionizing radiation.

Polyethylene exhibits good electrical and mechanical properties, moisture resistant and not soluble in many solvents except benzene and petroleum at high temperature. The dielectric constant and power factor remains steady over a wide range of temperature.

It is used as general purpose insulation, insulations of wires and cable conductors, in high frequency cables and television circuits, jacketing material of cables. Polyurethane films are also used as dielectric material in capacitors.

Teflon: The chemical name of Teflon is Polytetra fluoro-ethylene. This is synthesized by polymerization of tetra fluro ethylene. It bears good electrical, mechanical and thermal properties. Its dielectric constant is 2 to 2.2, which does not change with time, frequency and temperature. Its insulation resistance is very high and water resistant.

It is used as dielectric materials in capacitors, covering of conductors and cables, as base material for PCBs.

Polyvinyl Chloride (PVC): It is obtained by polymerization of vinyl chloride in the presence of a catalyst at 50^0C. PVC exhibits good electrical and mechanical properties. It is hard, brittle, and non-hygroscopic and can resist flame and sun light.

PVC used as insulation material for dry batteries, jacketing material for wires and cables.

Epoxy Glass: Epoxy glass is made by bonding two or more layer of material. The layers used reinforcing glass fibers impregnated with an epoxy rasin. It is water resistant and not affected by alkalis and acids.

It is used as base material for copper-clad sheets used for PCBs, terminal port, instrument case etc.

Bakelite: It is hard, dark colored thermosetting material, which is a type of phenol formaldehyde. It is widely used for manufacture of lamp holders, switches, plug socket and bases and small panel boards.

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Dielectric Material

Introduction:

The materials which are capable of retarding the flow of electricity or heat through them are known as dielectric or insulators. The safe handling of heat and electricity is almost impossible without use of an insulator. The material when used to prevent the loss of electrical energy and provides a safety in its operation is named as Electrical Insulating Material. The properties which are taken into consideration for an insulator are the operating temperature and breakdown voltage. However when it is used to store electrical charge, it is known as Dielectric Material.

The electrical conductivity of Dielectric material is quite low and the band gap energy is more than 3eV. This is the reason why the current cannot flow through them. The capacity of a capacitor can be increased by inserting with a dielectric material, which was discovered by Michael Faraday.

Dielectric Parameters:

The knowledge of dielectric parameter is highly essential to choose the specific purpose dielectric for use. Those are *Dielectric constant, Dipole moment, Polarization and Polarizability.*

- ***Dielectric constant:*** The proportionality constant in the relation between the electric flux density (D) and the electric field intensity (E) is known as permittivity (ε) or dielectric constant. If the medium to which the electric field is applied is a free space (or vacuum), the proportionality constant of vacuum is ε_0 of value 8.854 × 10^{-12} farad.meter^{-1}. The dielectric constant of a material may be expressed as ε_r, relative to that of a vacuum by, $\epsilon_r = \frac{\epsilon}{\epsilon_0}$. So the relation of electric flux density and electric field intensity is given by,

$$D = \varepsilon_0 \varepsilon_r E$$

  Where ε_r is a dimension less quantity and is known as relative dielectric constant, which is determined by the atomic structure of the material.

- **Dipole Moment:** Two charges (Q+ and Q-) of equal magnitude but of opposite polarity, separated with distance d, constitute a dipole moment , given as: **p** = Qd
  p is the dipole moment in coulomb-meter. Dipole moment is a vector pointing from the negative charge to the positive charge and its unit is Debye (1 Debye = 3.33 × 10^{30} coulomb-metre).

Dipole Moment

- **Polarization:** The dipole moment per unit volume is called the polarization **P**. $P = \frac{p}{volume}$; where p is the dipole moment and P is the polarization in coulomb.meter^{-3}.

  Considering a parallel plate capacitor having two metal plates of area A and separated in vacuum by distance d and having a battery of voltage V connected across it. The electric field E between the plates is given by V/d volt.m^{-1} arising from the charge density ±Q on the plates. The relation between Q and E is given by, $Q = \varepsilon_0 E$.

  Q can be considered as a source of electric flux lines in the space between the plates; the density of this flux lines is the electric displacement D.

  $$D = Q = \varepsilon_0 E.$$

  Now consider that the battery is still connected and a dielectric medium is introduced to fill the space between the plates. The medium becomes polarized by the field E and dipoles appear throughout the material, lined up in the direction of the field. All dipole ends of opposite charge inside the material will cancel, but there will be an uncompensated surface charge on the plates, Positive on one plate and the negative on the other plate. These surface charges will attract and hold corresponding charges of opposite sign on the plates because the latter, unlike dipoles are able to move freely. The field in the dielectric will be still E. If the effects of some of the original surface charges have been neutralized by being bound to surface dipole ends, E can only be maintained by the flow of more charges on the battery to compensate for those, which has become bound. There is now more charge density Q' on the plates some of which is tied up and is not contributing to the field E in the dielectric. The amount of charge that is contributing to the field is the same as before and $Q' = Q + Q_B$.

  Where Q_B is the bound charge density; Q has been multiplied by a factor ε_r such that $Q' = \varepsilon_r.Q$. Electric field density is now given by;

  $$D = \varepsilon_0 \varepsilon_r E$$

  $$or, D = \varepsilon_0 E + Q_B$$

  The bound charge density is called polarization P. This is identical with the dipole moment per unit volume.

  The polarization may be expressed in terms of elementary dipole moments p by,

  $$P = N.p ;$$

  Where N is the number of dipoles per unit volume.

- **Polarizability:** The application of an electric field to a dielectric material causes a displacement of electric charges giving rise to the creation or reorientation of the dipoles in the material. The average dipole moment 'p' of an elementary particle may be assumed to be proportional to electric field strength E, that acts on the particle so that; $p = \alpha E$

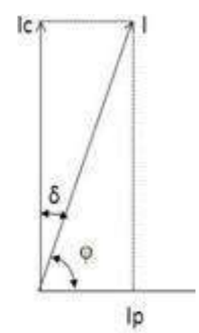
The proportionality factor α is called polirizability, measures the average dipole moment per unit field strength. The unit of the polarizability is farad.meter^2.

Mechanism of Polarization: The centre of gravity of positive charges and negative charges coincide in neutral atoms and symmetric molecules. When an electric field is applied to it, causes relative displacement of charges, leading to the creation of dipoles and hence polarization takes place. Un-symmetric arrangement of atom in a molecule results in a dipole even in the absence of an external field and in those cases the applied electric field tends to orient the dipole moments parallel to the field direction. The mechanism for forming the dipoles are categorized as (i) Electronic or Induced polarization, (ii) Ionic polarization, (c) Orientational polarization, (d) Interfacial or Space charge Polarization. Discussion of above mechanisms is restricted within the scope of the syllabus.

Dielectric Loss: The dielectric material separating the two electrodes or conductors is stressed when subject to a potential. When the potential is reversed, the stress also reversed. This change of stress involves molecularly arrangement within the dielectric. This involves the energy loss with each reversal. This is because the molecules have to overcome a certain amount of internal friction in the process of alignment. The energy expanded in the process is released as heat in the dielectric.

The loss appearing in the form of heat due to reversal of electric stresses, compelling molecular arrangement is known as dielectric loss.

When a dielectric material is subjected to an ac voltage, the leakage current I does not lead the applied voltage E by exactly 90^0. As shown in vector diagram the phase angle ϕ is always less than 90^0. The dielectric loss can be calculated as follows:



$$P = E\,I\,cos\emptyset$$

$$where \;\; \emptyset = 90^0 - \delta \; and \;\; I = \frac{I_c}{\cos\delta}$$

$$\therefore P = E\,\frac{I_c}{\cos\delta}\,.\cos 90^0 - \delta \;=\; E\,\frac{I_c}{\cos\delta}\,.\,sin\delta = E\,I_c tan\delta = E.\frac{E}{X_c}tan\delta$$

$$Hence, P = \;E^2\,2\pi fc.tan\delta$$

δ is the complement angle to ϕ and is called dielectric loss angle. tanδ is the measure of dielectric loss known as dissipation factor. Good dielectric material should have very small dissipation factor yo minimize dielectric loss.

Factors affecting Dielectric loss: As observed from the equation of dielectric loss, the loss depends on the frequency and square of applied voltage. Dielectric loss increases with the presence of humidity and temperature rise.

Electrical conductivity of Dielectric and their Breakdown

The dielectric material is used in electrical and electronic circuits as insulators and as a medium in capacitors. When the applied electric field is increased, the potential difference across it also increases. A limit is reached when the dielectric ceases to work as an insulator and a spark occurs. This limiting value of the voltage is known as Breakdown Voltage, which measures the strength of dielectric.

$$Therefore, dielctric\ strength = \frac{Breakdown\ voltage}{Thickness\ of\ the\ dielectric}$$

Conduction of Gaseous dielectric: Air is the common gaseous dielectric. Cosmic rays and Ultraviolet rays cause the natural ionization in air. Since the opposite charges are equal, natural recombination takes place continuously to check further ionization of whole air.

The free charges do not go for recombination if the medium is within an Electric field. Due to application of the electric field, free charges move to their respective potential plates, causing a flow of current known as leakage current. The magnitude of current is dependent upon the applied voltage. With the increase in voltage the directed flow of electrons and ions increases as compared to random motion in low voltage. If the applied voltage is further increased, the energy of free charges becomes sufficient to force out electrons even from neutral atom. Each free electron moves at a great velocity, collides with other neutral atoms and knocks out free electron out of them. This process increases in geometric progression. The leakage current increases sharply in result to cause the breakdown of dielectric. The corresponding voltage is known as Breakdown voltage.

Conduction of Liquid dielectric: The liquid dielectric along with impurities of solid particle has more ability to conduct. The impurities get electrically charged and act as a current carrier. The fibrous impurities make the alignment of ions in a straight path for which the conductivity in liquid gets faster. In an uncontaminated liquid dielectric, such ion bridge cannot be formed. The breakdown of an uncontaminated liquid dielectric takes place due to the ionization of gases present in the liquid. The applied voltage ionizes the gas in liquid and the electric field intensity increases. It causes further ionization and ultimately the breakdown of dielectric takes place.

Conduction of solid dielectric: Electrical conductivity of solid dielectrics may be electronic, ionic or both. In electronics current flow the flow of current is due to the movement of electrons towards the positive electrodes, while ionic current flow is due to the movement of positively charged ions towards the negative electrode. The impurities also play the role of conductivity in the dielectric. At low temperatures, the conductivity of solid dielectric is due to the impurities only. At higher temperature the leakage current depends upon the contribution of free ions of the base dielectric.

Breakdown of solid dielectrics may be electro-thermal or electrical. The heat produced due to dielectric loss causes electro thermal breakdown and in effect destruction of dielectric takes place. If the dielectric is not able to radiate away the generated heat caused by dielectric loss and the applied voltage is retained for a long period the material gets melted. The electrodes get short circuited. Solid dielectric is not recoverable after its break down like liquid or gaseous dielectrics.

Properties of Dielectric Materials:

Some of the main properties of important dielectrics used in practice are given in following table:

Material	Dielectric constant	Dielectric strength (kV/mm)	tan δ	Max working temp at 0^0C	Thermal conductivity (mW/mK)	Relative density
Air	1	3	-	-	0.025	0.0013
Alcohol	2.6	-	-	-	180	0.79
Asbestos	2	2	-	400	80	3.0
Cellulose film	5.8	28	-	-	-	0.08
Cotton fabric (dry)	-	0.5	-	95	80	-
Impregnated	-	2.0	-	95	250	-
Ebonite	2.8	50	0.005	80	150	14
Glass (flint)	6.6	6	-	-	1100	4.5
Glass (crown)	4.8	6	0.02	-	600	2.2
Mica	6	40	0.02	750	600	2.8

Dry Paper	2.2	5	0.007	19	130	0.82
Impregnated paper	3.2	15	0.06	90	140	1.1
Quartz	5.7	15	0.008	1000	1000	2.4
Vulcanized Rubber	4	10	0.01	70	250	1.5
Resin	3	-	-	-	-	1.1
Fused Silica	3.6	14	_	-	-	-
Silk	-	-	-	95	60	1.2
Sulphur	4	-	0.0003	100	220	2.0
Water	7.0	-	-	-	570	1.0
Paraffin Wax	2.2	12	0.0003	35	270	0.88

Application of Dielectrics:

The most common application of dielectric is as a capacitor to store energy. Capacitors are classified according to use of dielectrics used in their manufacture.

i. Capacitors using vacuum, air or gases as dielectric.
ii. Capacitors using mineral oil as dielectric.
iii. Capacitors using a combination of solid and liquid dielectrics.
iv. Capacitors only with solid dielectrics like glass and mica etc.

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DIELECTRIC LOSS:-

Consider a parallel plate condenser filled with dielectric material characterised by r ’ . Let the electrode area be ‘A’ and the plate separation be ”d”. The admittance of the capacitor for any angular frequency ω is given by =G+jB where G and B are the conductance and susceptance respectively.

Where C is the capacitance which may depend on frequency, because C=r ’ A/d. The conductance G equals σA/d, where σ is an effective conductivity at the angular frequency ω. The conductance G arises because of the conversion of part of the electrical energy into heat, but mainly because of the complex dielectric constant.

Thus the absorption of energy by the material in an alternating field is proportional to the imaginary part of the dielectric constant. The dielectric is said to have losses which are characterised by the loss tangent.

In this circuit Īc is the conduction current which is responsible for the dielectric losses and is in phase with the applied voltage. Īd is the displacement current which is in phase quadrature with the applied voltage. Ī is the phasor sum of Īc and Īd. If there are no losses, r ” is zero and the current Ī leads the applied voltage by 900 under these circumstances.

ELECTRIC CONDUCTIVITY OF DIELECTRICS AND THERE BREAK DOWN:-

Electrical breakdown or dielectric breakdown is a process that occurs when an electrical insulating material, subjected to a high enough voltage, suddenly becomes an electrical conductor and electric current flows through it.

The electric strength at breakdown is defined as the minimum electric stress usually expressed in KV/cm, waveform, frequency and the type of electrodes. The electric breakdown strength of a material depends on its composition, thickness, temperature, moisture content and to some extent on the time of application voltage. It is also affected by the shape of the waveform and the steepness of the waveform of the applied voltage. There is no definite relationship between these variables, but in general for sheet materials, the electric strength is an inverse function of the thickness and time and decreases with increases temp. And moisture content. At breakdown, the high electric stress is assumed to cause an interatomic displacement of the orbital electrons which alters the atomic structure causing heating and a conducting path in the material. The breakdown mechanism of gaseous, liquid and solid dielectrics are different in nature.

Gaseous Dielectrics:-

Breakdown in gases begins with the ionization due to collision of electron .In strong electric field The kinetic energy acquired by the accelerating free electron will be greater than the ionization energy of the gas. Collision ionization is started by the most of the mobile electron. The dielectric strength of gaseous dielectric depend on many factor. The electric strength is higher at very high and at very low pressure. The electric strength also depend on the uniformity of the applied electric field. When the electrode have different shapes the breakdown voltage depend on the polarity of the electrodes. The breakdown voltage depends on the frequency of the applied field. The other factors which influence the breakdown voltage of gas dielectric are the distance between the electrode and the chemical composition of the gas.

Liquid Dielectric:-

The breakdown mechanism of liquid dielectric depend on the purity of the dielectric. In the contaminated dielectric the breakdown occur due to the formation of conducting bridge between the electrode by droplet of emulsified water and suspended particles especially fibrous particles. The time taken to form the bridge depends on the extent of combination, the shape of the electrodes and gap between them. In technically pure liquid dielectric the break down is initiated by the ionisation of the gases in the liquid. All liquids dissolve a certain quantity of gas especially air. In degassed high purity liquid dielectric breakdown is evidently due to the collision ionization initiated by secondary electron emitted from the cathode due to the strong electric field.

PROPERTIES OF DIELECTRIC:-

Following are the exhibited by the dielectric materials:

- The energy gap in the dielectric materials is very large.
- The temperature coefficient of resistance is negative and the insulation resistance is high.
- The dielectric materials have high resistivity.
- The attraction between the electrons and the parent nucleus is very strong.
- The electrical conductivity of these materials is very low as there are no free electrons to carry current.

APPLICATION OF DIELECTRIC:-

Some of the applications of dielectrics are as follows-

- These are used for energy storage in capacitors.
- To enhance the performance of a semiconductor device, high permittivity dielectric materials are used.
- Dielectrics are used in Liquid Crystal Displays.
- Ceramic dielectric is used in Dielectric Resonator Oscillator.
- Barium Strontium Titanate thin films are dielectric which are used in microwave tunable devices providing high tunability and low leakage current.
- Parylene is used in industrial coatings acts as a barrier between the substrate and the external environment.
- In electrical transformers, mineral oils are used as a liquid dielectric and they assist in the cooling process.
- Castor oil is used in high-voltage capacitors to increase its capacitance value.

Electrets, a specially processed dielectric material acts as electrostatic equivalent to magnets

MAGNETIC MATERIAL

(CHAPTER 5)

- Materials which can be magnetised are called magnetic material
- When magnetised, such material creates a magnetic field around them.
- When current flows through a coil it creates an MMF. Does the circulating electron in a material also develop MMF In most material the direction of motion of the electron in various orbit is such that they develop MMF In opposite direction thus cancelling each other.
- In magnetic material there are a number ofunneutralised orbits which produce a resultant MMF creating magnetic poles called magnetic dipoles.
- In an unmagnetized material the Dipoles are scattered at random.
- In a magnetic material the dipole line of parallel with exciting MMF.
- The property of a material by virtue of which it allows itself to magnetised is called permeability.
- In magnetic material, the value of permeability is constant and is the same as for the free space.
- The permeability of free space is denoted by R_o and equals to $7\pi\times 10^{-7}$
- The permeability of air is almost same as for the free space i.e, $7\pi\times 10^{-7}$
- For magnetic material the permeability μ is given by $\mu = \mu_o \times \mu_r$. Where μ_r is called the relative Permeability.
  $B = \mu_o \times H$
- Relative permeability depends upon the nature of the material and on temperature.
- When relative permeability is positive the magnetic dipoles arranged themselves in the same direction as the applied field intensity.
- When relative permeability is negative the magnetic dipole is allign themselves in the opposite direction to the applied field.

The material are classified into:

1. **Magnetic material:**
   These material respond to an external magnetic field.
2. **Non-magnetic materials:**
   These materials Do not respond to an external magnetic field.

Magnetic metals are classified in:

1. Diamagnetism
2. Para magnetism
3. Ferromagnetism

- This classification is made on the manner they respond to external magnetic field.
- Diamagnetic and paramagnetic falls into the category of non-magnetic Materials.
- Ferromagnetic materials are classified as magnetic.

DIAMAGNETISM:

- Permanent magnetic dipole are absent in them because there is cancellation of magnetic fields due to electron rotating in opposite direction in the various orbit of the atom is total.
- Materials which lack permanent magnetic dipoles are called diamagnetic material.
- If an external magnetic field is applied to a diamagnetic materials it introduce a magnetism (M) in opposite direction to the applied field intensity (H).
- The relative permeability μ_r , of diamagnetic material is negative due to this reason diamagnetic material are not used in electrical engineering application.

For example: Silver, copper, bismuth, Hydrogen, etc.

PARAMAGNETISM:

- Many metals have small but positive relative permeability such material are called paramagnetic.
- The atomic dipoles are oriented in a random fashion.
- The resultant magnetic field is thus Negligible
- On application of an external magnetic field the permanent magnetic dipoles orient themselves parallel to the applied magnetic field and give rise to a positive magnetisation (M).
- The orientation of Dipole is parallel to the applied magnetic field is not complete therefore the magnetism (M) is small. The relative permeability of paramagnetic material are approximately unity. Thus paramagnetic material have negligible Application in the Field of electrical engineering.

For example: aluminium, platinum, oxygen, etc.

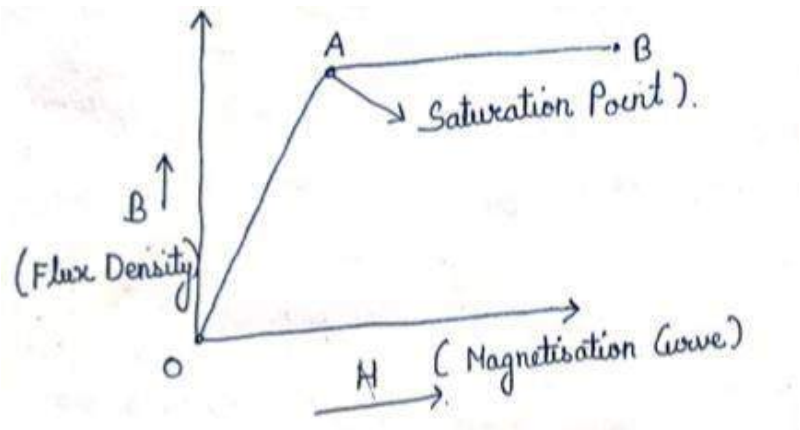
FERROMAGNETISM:

- Ferromagnetic material crystalline solids. The permanent atomic dipoles are aligned parallel to each other within group called Domains. Each domain is therefore at all times completely magnetised. When a weak external magnetic field is applied, it is not enough to cause any change in the orientation of dominance. The flux density with search low applied field is entirely Due to the externally applied magnetic field.
- When the externally applied Magnetic field is increased, stage where domain is still weak, it will start or renting themselves search that their resultant magnetic field coincides with the externally applied magnetic field and the material Will develop a strong magnetic field of it's own.
- There are some domains Who is original magnetic orientation greatly diverges from that of the applied Field and required a stronger external field to be Able to orient their magnetism in the same Direction as the applied field. i.e, determines whose original direction of magnetism is less divergent from that of the applied field, the rate

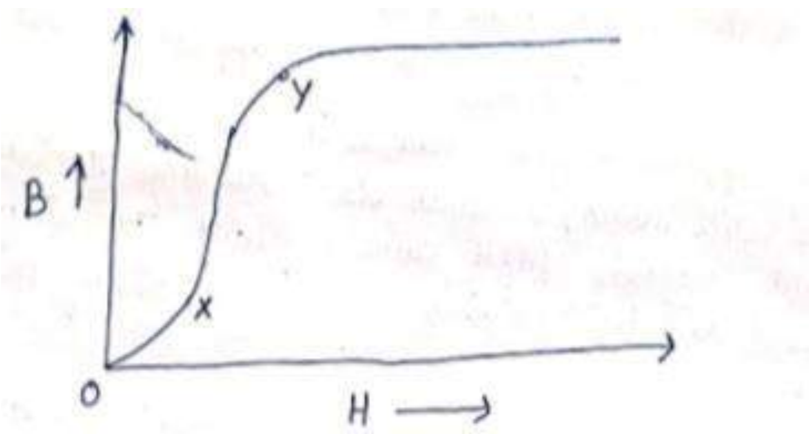
of strengthening of the internal magnetic field decrease with increase in the applied magnetic field and gives rise to magnetic saturation.

For example: Iron, cobalt, Nikel, etc

MAGNETISATION CURVE: -



OR

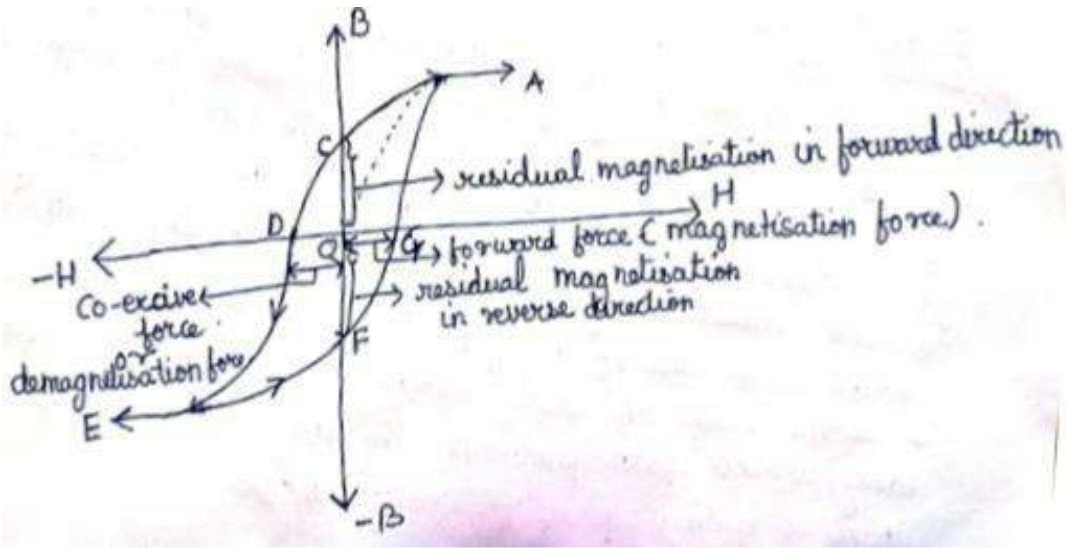


With very weak external field H, the flux density B, rise in direct proportion (i.e. as a straight line from origin).This means that during this reason up to the point X, the domains of the ferromagnetic material do not Orient themselves. Parallel to the applied field and therefore the material is not magnetized. The flux density is entirely due to the external field thus permeability of the material up to the point X is constant is called initial permeability. if the external field H is increased beyond the point X ,there is sharp increase in the flux density because the external field is strong enough to Orient parallel to its own axis. Up to point y the relative permeability of the material is not constant but keeps increasing.

when the magnetization curve reaches the point why the material is start saturating. After the curve becomes almost zero.

HYSTERESIS: -

- When ferromagnetic material the flux density B increases when external magnetic field applied to it is increased,when saturation arrives the increase in B almost cases even though H may be increased.
- If the external magnetic field is reduced it is found that the curve OA is not retraced.At H=0,the material is still magnetised and the flux density B has the value OC which is called residual magnetism.
- In order to demagnetize the material completely the external magnetic field H, is reversed when H reaches the value OD in the rivers to direction and BO in this time.Hence this applied magnetising force H in the reverse direction which causes B to be zero is co-ercive force. This force is also known as demagnetisation force.
- Further increase of H in the reverse direction will now increase B in the reverse direction and again at point E saturation occur.
- Again the external field is increased in forward direction and when H=0,her B=0 which is called the residual magnetisation.
- To neutralize the residual magnetization H is increased to the value of OG on the positive i.e the original direction. Further increase of H in the positive direction will again magnetise the material in this direction and saturation will occur at A.



- Flux density B always lags behind H
- this property of B lagging behind H is a characteristic of the magnetic behaviour of the ferromagnetic materials. When H is taken from positive maximum through zero to negative maximum and then through zero again back to positive maximum ,the graph relating B&H tracer loop ACDEFGA.this is called Hysteresis loop.

- when a ferromagnetic material is subjected to repeated cycles of magnetisationa loss occur this losses due to hysteresis is known as hysteresis loss this loss is directly proportional to the supply frequency.
- Hysteresis loss depends upon flux density and frequency of variation of flux and can be expressed as,

  Hysteresis loss=K $B_m^{1.6}$ f V $_c$watts

  Where K is a constant whose value depend upon the core material B_m is the maximum flux density of the magnetic field in which the core is placed.
- Magnetic core used in transformer and rotating electrical machines are made from materials whose hysteresis loops are narrow in order to keep down hysteresis loss.

EDDY CURRENT: -

Magnetic materials placed in alternating magnetic fields also have eddy currents induced in them. This is because the material is subjected to rate of change of flux linkage according to faradays law of electromagnetic induction E.M.F are induced in the material causing currents, called eddy currents, to flow in the material. These current causes loss of energy(I^2R)loss in the material, where I is the value of eddy current and R is the resistance of the eddy current path provided by the material. Due to the loss there is heating up of the material.

- The eddy current loss is proportional to the square Of frequency, square of the thickness of the material and inversely proportional to the resistivity of the material.
- To prevent Eddy current loss magnetic cores to be used in alternating magnetic field are built up on thin sheets of steel lamination separated from each other by thin film of insulation.
- The insulating film may be a cot of insulating varnish,a sheet of a paper or a film of oxide .
- An efficient insulation for silicon steel is a film of phosphate chemically deposited on the surface this film hold the high temperature annealing required by lamination punched out of sheets .This increases the resistance to the path of eddy current thus reducing their magnitude and eddy current loss.
- The losses due to hysteresis and Eddy current are also affected by the magnitude of the flux density hysteresis loss varies directly in proportional to B at low flux densities say upto 0.1Wb/m^2 to 1.0Wb/m^2 it is proportional to $B^{1.6}$.
- Eddy current losses are proportional to the square of flux density.

Eddy current loss=$KB^2mf^2t^2V_c$ watt

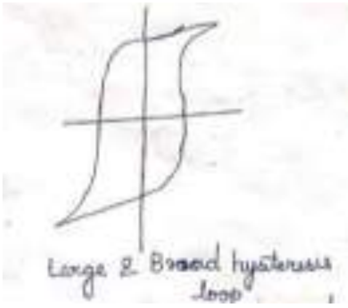
Where K is a constant which depend upon the core material amd t is thickness of the core lamination.

CURIE POINT:- The critical temperature above which the ferromagnetic materials losses their magnetic properties is called curie point.

At Curie point or Curie temperature the domain structure tend to disrupt and the domains lose their alignment, become arranged in random fashion thus the material loses its ferromagnetic property.

MAGNETOSTRICTION:- When ferromagnetic material are magnetised a small change of dimension of the material takes place there is a small extension with reduction of cross section of the crystal of which the material is made when subjected to rapid alternating magnetic field there is a rapid and continuous extension and contraction of the material. This is called magnetostriction.

All ferromagnetic materials are divided into two groups:

SOFT MAGNETIC MATERIAL	**HARD MAGNETIC MATERIAL**
● Materials which has a steeply rising magnetization curve ,relatively small and narrow hysteresis loop and small energy losses during cyclic magnetisation are called soft magnetic material.  ● Used in building cores for use in alternating magnetic fields. ● Eg-soft pure iron,siliconiron,alloy,nickel iron alloy and soft ferrites. ● Used in construction of cores in electric-machine,transformers,electro-magnets,reactors,relays etc. ● They have small enclosed area of hysteresis loop. ● High permeability. ● High saturation value of flux.	● Magnetic materials which have a gradually rising magnetization curve large hysteresis loop and large energy losses during cyclic magnetisation are called hard magnetic material.  ● Used for making permanent magnets. ● Eg-carbon steel,tungstensteel,cobaltsteel,alnico, hard ferrites ● They have high saturation values,highcoersive force and high residual magnetisation.

Example of soft magnetic material:-

1. Pure iron:-It is a ferrous with an extra lower carbon content.eg-low carbon steel and electrolyte iron.in low carbon steel the carbon content is less than 0.1%.it has high magnetic permeability.

The resistivity of pure iron is very low by virtue of which it gives rise to large eddy current losses when operated at high flux densities in alternating magnetic fields.

Pure iron is widely used in electrical apparatus and instrument as magnetic material care for electromagnets,components for relay electrical instruments etc.

Pure iron is not used in rotating electrical machine where rotation of the slotted rotor cause variation in reluctance .

2. Iron silicon alloy:-The application of pure iron is limited because of its high eddy current losses.so by adding 0.5to 5%by weight of silicon in to iron is known as iron silicon alloy usually called silicon steel.

It is used for strong alternating magnetic fields generally used in transformer,electrical rotating machine,reactors,electromagnets and relays.

Silicon sharply increase the electrical resistivity of iron thus decreasing the iron losses due to eddy current .it increase the permeability at low and moderate flux densities but decreases it at high densities.

Addition of silicon to iron reduces the hysteresis loss. The magnetostriction effect is also reduced.

By addition of silicon into iron its improved the magnetic properties of iron and fasilates the steel making process.

Alloying of carbon steel with silicon increases the tensile strength, it reduces ductility mobbing steel brittle.

The silicon alloyed steel difficult to perch and sheathe .a steel with more than 5%silicon may be too hard and brittle.

The care are made up silicon alloy to reduce the hysteresis loss and care is laminated to minimum the eddy current loss.

GRAIN ORIENTED SHEET STEEL:-

Grain-oriented electrical steel usually has a silicon level of 3% (Si:11Fe). It is processed in such a way that the optimal properties are developed in the rolling direction, due to a tight control (proposed by Norman P. Goss) of the crystal orientation relative to the sheet.
The magnetic flux density is increased by 30% in the coil rolling direction, although

its magnetic saturation is decreased by 5%. It is used for the cores of power and distribution transformers, cold-rolled grain-oriented steel is often abbreviated to CRGO.

CRGO is usually supplied by the producing mills in coil form and has to be cut into "laminations", which are then used to form a transformer core, which is an integral part of any transformer.

NICKEL AND IRON ALLOYS:-

• Iron and iron nickel alloys have low critical permeability. For power application like transformers and rotating electrical machines where core materials are operated at high flame densities the initial permeability is of no importance. Put for high sensitivity and low distortion needed in communication systems. The iron and silicate alloys are not suitable.

• In special magnetic alloys having initial and maximum permeability are used for special application like instrument transformers, relays etc. A group of iron alloys containing nickel between 30 to 80 percent with possible selection is molybdenum and chromium when given appropriate treatment during manufacture, show very high permeability's at low flux densities and much lower losses then iron. The important alloys in this category are permalloy , super alloy and mumetal. Permalloy: This is used in manufacture of sensitive relays. The curie tempertature of this group of material varies between 420 degrees to 580 degrees depending on the percentage of nickel content and heat treated upto about 1050to 1100 degrees celisus. It has initial permeability between 2500 to 8000 as well as maximum permeabilities is 1,00,000. Superalloy: It consists of iron and nickel alloyed with copper and molybdenum. This alloy has high initial permeability upto 1,00,000. Mumetal: It consist of iron and nickel alloyed with copper and chromium. It is used for manufacture instrument transformer and miniature transformer used in communication circuits. Its initial permeability is 20000 and maximum permeability is 110,000. Its curie temperature is 130 degree Celsius.

• **Super ferrites**: Ceramic magnets, also called ferromagnetic ceramics and ferrites are made of Iron Oxide Fe2O3 with one or more divalent oxides such as NiO,MnO,ZnO.This magnets have a square hysteris loop and high resistanc to demagnetization and are used for computing machines where high resistance is required.They have high resistivity and reduced eddy current. Ferrites are made by mixing powdered oxides,compacting and sintering at high temperature.Ferrites cores are used in high frequency transformer in television and frequency modulated receiver. Ferrites with large magnetostriction effect are sometimes used in electromechanical transducer. In high frequency application,magnetostruction in ferrites can lead to undesirable noise.

Examples of Hard magnetic material:

Hard magnetic materials are made up of permanent magnate

1. Carbon steel, tungsten steel and cobalt steel :Soft magnetic material cannot be used for making permanent magnets because they have narrow hysteresis loops. when Carbon is added in a material, the hysteresis loop area is increased i.e. Carbon steel was used for permanent magnet. It is cheap, magnets made from carbon steel lose their magnetic properties very fast sunder the influence of knocks and vibration.

When materials like tungsten, chromium or cobalt are added to carbon steel, its magnetic properties improved. Cobalt steel has superior magnetic properties but it is expensive.

2. Alnico: Alloys like ALNICO(Aluminium-nickel iron cobalt) are hard magnetic materials.They are hard magnets used in many electrical engineering applications i.e. in various measuring instruments.

3. Hard ferrites: Hard magnetic ferrites like $BaO(Fe_2O_3)_6$ are used for manufacture of light weight permanent magnets due to low specific weight.

MATERIALS FOR SPECIAL PURPOSES

Some materials used for special purposes such as fuses, solder, bimetal, storage battery plates. Those materials used for special purposes are in structural materials or protective materials.

STRUCTURAL MATERIALS :

Cast iron, steel, timber, reinforced concrete are common materials for this purpose.

Cast iron is used as materials for the frames of small and medium sized electrical machines. Steel is used in fabricated frames in large electrical machine, tanks in a transformers, fabrication of transmission towers.

Timber and reinforced concrete are used for poles in OH lines.

PROTECTIVE MATERIALS :

LEAD :

Lead is soft, heavy and bluish grey metal. It is highly resistant to many chemical action, but it can corrode by nitric acid, acetic acid, line and rotten organic substance. The electrical conductivity is 7.8% of copper. Lead is used in storage batteries and sheathing of cables. Pure lead cable sheathing are liable to fail in service due to formation of cracks formed because of vibration.

Lead alloys with tin and zinc and forms alloys which are used for solders and bearing metals.

STEEL TAPES, WIRES AND STRIPS :

Steel tapes, wires and strips are used as protective materials for mining cables, underground cable, weather proof cables.

OTHER MATERIALS :

THERMO COUPLE MATERIALS :

When two wires of different metals are joined together an emf exist across the junction. This emf is directly proportional to the temperature of the junction. When one tries to measure this emf more junctions are to be made which will give rise to emfs. When all the junctions are at the same temperature, the resultant emf will not be zero. This resultant emf is proportional to the temperature difference of the junctions and is known as thermoelectric emf.

Thermo couples are made of different materials such that copper / constantan, iron / constantan, platinum / platinum rhodium.

Thermo couples can be used for the measurement of temperature.

BIMETALS :

A bimetal is made of two metallic strips of unlike metal alloys with different coefficient of thermal expansion. At a certain temperature the strip will bend and actuate a switch or a lever of a switch. The bimetal can be heated directly or indirectly. When heated the element bends so that the metal with the greater coefficient of expansion is on the outside the are formed while that with smaller coefficient is on the inside.

Bimetallic strips are used in electrical apparatus and such as relays and regulators.

SOLDERING MATERIALS :

An alloy of two or more metals of low melting point used for base metals is known as soldering. The alloy used for joining the metals is known as solder. The solder is composed of 50% lead and 50% tin. Its melting point is $185^{o}C$ tensile strength is 385 kg./cm^2 and electrical conductivity is 10% of copper.

For proper soldering flux is to be used. In soldering process the application of flux serves to removes oxides from the surface to be soldered. They deoxidize the metals at the time the soldering element is added. Solders are two types such as soft solders and hard solders soft solders are composed of lead and tin in various proportions. Hard solders may be any solder with a melting point above that of lead tin solders.

The application of soft solders is in electronic devices and hard solder in power apparatus for making permanent connection.

EYRE NO.7 FLUX :

It is an improved variety of organic flux which is used with Alca P for alumnimum cable jointing. This on decomposition at a temperature a little below the jointing temp approx $316^{o}C$ removes the refractory oxide from the strands of the core and makes the surface receptable to solder.

FUSE :

A fuse is a protective device, which consists of a thin wire or strip. This wire is placed with the circuit which heave to protect, so that the circuit. Current flows through it. When this current is too high the temperature of the wire or strip will increase till the wire or strip melts. So braking the circuit and interrupting the power supply.

FUSE MATERIAL :

A fuse material have following properties :

a. Low resistivity
b. Low conductivity
c. Low melting point

As lead is used as fuse material because of its low melting point. But the resistivity of lead is high, thick wires are used. For rewirable fuses alloys of tin and lead are used.

DEHYDRATING MATERIAL :

SILICA GEL :

It is an in organic chemical, a colloidal highly absorbent silica used as a de-humidifying and dehydrating agent as a catalyst carrier. Calcium chloride and silica gel are used in dehydrating breathers to remove moisture from the air entering a transformer as it breathes. Now silica gel is used for breather of a transformer. Its main advantage is that when it becomes saturated with moisture it does not restrict breathing. Silicagel when dry is blue in colour and the colour changes to pale pink as it becomes saturated with moisture.

LECTURE NOTES

ON

ENTREPRENEURSHIP AND MANAGEMENT & SMART TECHNOLOGY

COMMON TO ALL BRANCH

5TH SEMESTER

UNIT-1
ENTREPRENEURSHIP

- Concept /Meaning of Entrepreneurship
- Need of Entrepreneurship
- Characteristics, Qualities and Types of entrepreneur, Functions
- Barriers in entrepreneurship
- Entrepreneurs vrs. Manager
- Forms of Business Ownership: Sole proprietorship, partnership forms and others
- Types of Industries, Concept of Start-ups Entrepreneurial support agencies at National, State, District Level(Sources): DIC,NSIC,OSIC, SIDBI, NABARD, Commercial Banks, KVIC etc.
- Technology Business Incubators (TBI) and Science and Technology Entrepreneur Parks

Concept /Meaning of Entrepreneurship

The word entrepreneur comes from the 13th century from a French word "entreprendre" meaning to do something. The concept of entrepreneurship marks its establishment in the 1700s. Entrepreneurship is the ability and readiness to develop, organize and run a business enterprise, along with any of its uncertainties in order to make a profit. The most prominent example of entrepreneurship is the starting of new businesses.

An entrepreneur is someone who has an idea and who works to create a product or service that people will buy, as well as an organization to support that effort. An entrepreneur takes on most of the risk and initiative for their new business, and is often seen as a visionary or innovator.

Meaning of Entrepreneur

The entrepreneur is defined as someone who has the ability and desire to establish, administer and succeed in a start-up venture along with risk entitled to it, to make profits. The best example of entrepreneurship is the starting of a new business venture. The entrepreneurs are often known as a source of new ideas or innovators, and bring new ideas in the market by replacing old with a new invention.

It can be classified into small or home business to multinational companies. In economics, the profits that an entrepreneur makes are with a combination of land, natural resources, labour and capital.

***(i)* According to Peter F. Drucker** “Entrepreneurship is defined as a systematic innovation, which consists in the purposeful and organized search for changes, and it is the systematic analysis of the opportunities such changes might offer for economic and social innovation”.

***(ii)* According to Ricardo Cantillon** “Entrepreneurship entails bearing the risk of buying at a certain price and selling at uncertain prices.”

***(iii)* In the words of Joseph A. Schumpeter** “Entrepreneurship is any kind of innovative function that could have a bearing on the welfare of an entrepreneur.”

***(iv)* According to Robert K. Lamb** “Entrepreneurship is that form of social decision making performed by economic innovators.”

Need of Entrepreneurship

Creation of Employment- Entrepreneurship generates employment. It provides an entry-level job, required for gaining experience and training for unskilled workers.

Innovation- It is the hub of innovation that provides new product ventures, market, technology and quality of goods, etc., and increase the standard of living of people.

Impact on Society and Community Development- A society becomes greater if the employment base is large and diversified. It brings about changes in society and promotes facilities like higher expenditure on education, better sanitation, fewer slums, a higher level of homeownership. Therefore, entrepreneurship assists the organisation towards a more stable and high quality of community life.

Increase Standard of Living- Entrepreneurship helps to improve the standard of living of a person by increasing the income. The standard of living means, increase in the consumption of various goods and services by a household for a particular period.

Supports research and development- New products and services need to be researched and tested before launching in the market. Therefore, an entrepreneur also dispenses finance for research and development with research institutions and universities. This promotes research, general construction, and development in the economy.

Characteristics, Qualities and Types of entrepreneur

Not all entrepreneurs are successful; there are definite characteristics that make entrepreneurship successful. A few of them are mentioned below:

- **Ability to take a risk-** Starting any new venture involves a considerable amount of failure risk. Therefore, an entrepreneur needs to be courageous and able to evaluate and take risks, which is an essential part of being an entrepreneur.
- **Innovation-** It should be highly innovative to generate new ideas, start a company and earn profits out of it. Change can be the launching of a new product that is new to the market or a process that does the same thing but in a more efficient and economical way.
- **Visionary and Leadership quality-** To be successful, the entrepreneur should have a clear vision of his new venture. However, to turn the idea into reality, a lot of resources and employees are required. Here, leadership quality is paramount because leaders impart and guide their employees towards the right path of success.
- **Open-Minded-** In a business, every circumstance can be an opportunity and used for the benefit of a company. For example, Paytm recognised the gravity of demonetization and acknowledged the need for online transactions would be more, so it utilised the situation and expanded massively during this time.
- **Flexible-** An entrepreneur should be flexible and open to change according to the situation. To be on the top, a businessperson should be equipped to embrace change in a product and service, as and when needed.
- **Know your Product-**A company owner should know the product offerings and also be aware of the latest trend in the market. It is essential to know if the available product or service meets the demands of the current market, or whether it is time to tweak it a little. Being able to be accountable and then alter as needed is a vital part of entrepreneurship.

The most prominent entrepreneurs of past and present have had certain personal qualities.

1. **Willpower** is key to success. it is a strong longing to achieve success. Despite many failures, the entrepreneur never loses their will to achieve.

2. **Management** is the ability to make plans, devise targets, and then execute!

3. **Obsession -** The entrepreneur who is obsessed with their goals and ultimate vision will make others believe in them too.

4. **Smarts -** An entrepreneur must have old-fashioned common sense and be able to make good decisions and deal competently with others.

5. **Self-assurance -** The entrepreneur is self-assured, confident in their abilities and knows how to look into the future and plan well. A self-assured person instills further confidence in themselves.

6. **Devotion -** The entrepreneur is devoted to their goals, working long and hard, day and night to achieve them.

7. **Inspiration -** Building a business requires an inspiration and the creativity which will lead to a better product and excellent results,. An entrepreneur thinks outside of the box.

8. **Flexibility -** Creating a product that meets the needs of the market requires flexibility. This quality is very important. Stubborn, unflexible people will not make good entrepreneurs

Qualities of an Entrepreneur

1. Motivation
2. Creativity
3. Persuasiveness
4. Vision
5. Versatility
6. Risk Tolerance
7. Flexibility
8. Decisiveness
9. Collaboration

1. Motivation

Hard-working business owners are incredibly motivated to succeed. Adopting this mindset—and being able to demonstrate your motivation to an employer—is crucial. You need to bring enthusiasm to everything you do at your job.

Fortunately, showing you're highly motivated is simple: Show up to work every day with a positive attitude. This is among the top characteristics of entrepreneurship. Employers want to see you're passionate about what you do day in and day out.

2. Creativity

No matter what industry you're in, employers want workers with out-of-the-box ideas. They want to hire people to not only carry out assignments, but to also come up with better ways of doing things.

That's why it's important to be creative—to always be thinking of new ways you can improve your company's workflow, productivity, and bottom line.

3. Persuasiveness

Persuasiveness can make you a better negotiator, which gives you an edge when going after a plum assignment, raise, or promotion.

There will come a time when you need to convince a client, a coworker, or your boss to take certain actions, so you need to be persuasive when presenting your ideas.

4. Vision

Successful entrepreneurs always keep one eye on the big picture, and this ability can make you a better employee. Vision is primarily about strategic planning.

Can you see what direction the industry is going? Can you identify challenges for your company? Can you tackle your day-to-day job responsibilities, while staying focused on long-term goals and initiatives?

5. Versatility

Although you were hired for a specific set of skills, it's important that you can shift as needed. You want to be someone that your boss can go to in a pinch, so be prepared to tackle work that's outside your job description. Among the leading characteristics of entrepreneurship is being an early adopter of new technology and keeping your skills current.

6. Risk Tolerance

Every employer wants to grow their business, and that involves embracing change. Translation: Don't be afraid to take risks when pursuing new clients, for example, or testing a new product. (One caveat: Make sure you have your boss' buy-in.)

7. Flexibility

Like an entrepreneur, you have to be able to adapt to change and solve problems as they arise. A good team player can shift their priorities to help out whenever the team needs assistance. Thus, flexibility means being receptive to other people's needs, opinions, and ideas and being open-minded to feedback from your manager.

8. Decisiveness

Do you exercise sound judgment under pressure? When you're an entrepreneur, you don't have room to procrastinate—and the same is true for employees. You have to be comfortable taking action when needed. This means knowing how to prioritize tasks and make decisions quickly. (It helps to be organized.)

9. Collaboration

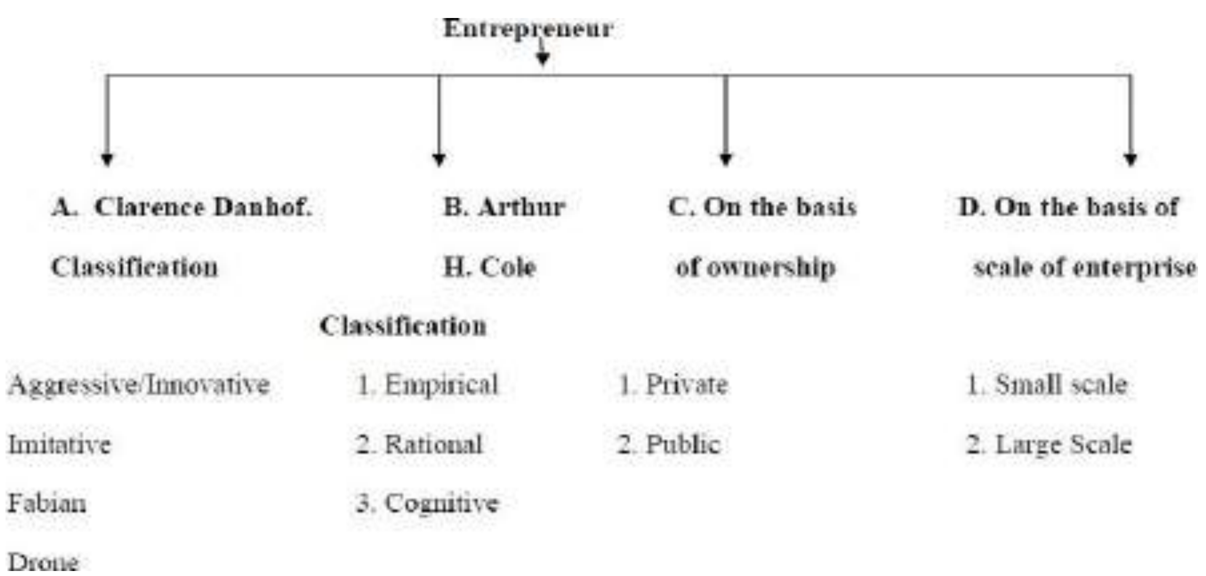
Savvy entrepreneurs are not only brilliant leaders, but also great collaborators, so you have to be an effective team player. Unsurprising, 86.3% of hiring managers seek job candidates who demonstrate strong teamwork skills, according to the National Association of Colleges and Employers Job Outlook survey.

Ultimately, using entrepreneurial skills at work entails adjusting to other people's work styles, avoiding office politics, celebrating your peers' successes, meeting your deadlines, and putting your company's goals first.

Types of entrepreneur

TYPES/CLASSIFICATION OF ENTREPRENEURS

Following are the classification of entrepreneurs on the basis of common characteristics.



A.) Clarence Danhof Classification:

In a study of American Agriculture, Based on their working relationship with the business environment they are functioning in, various types of entrepreneurs can be found, Clarence Danhof classifies entrepreneurs into four types-innovative, imitative, Fabian and drone.

1. **Innovating entrepreneurs**: This type of an entrepreneur is more interested in introducing some new ideas into the market, organization or in the nation. They are drawn towards innovations and invest a lot of time and wealth in doing research and development.

Innovative entrepreneurship is characterised by aggressive assemblage of information and the analysis of results derived from sound combination of factors. Persons of this type are generally aggressive in experimentation and cleverly put attractive possibilities into practice. An innovating entrepreneur sees the opportunity for introducing a new technique or a new product or a new market. He or she may raise money to launch an enterprise, assemble the various factors, choose top executives
and set the organisation going. Schumpeter's entrepreneur was of this type.
Such an entrepreneur introduces new products and new methods of production,
opens new markets and re-organises the enterprise.
Among the different types of entrepreneurs, the innovating entrepreneur is the most vigorous type of entrepreneur. Innovating entrepreneurs are very commonly found in developed countries. There is dearth of such entrepreneurs in underdeveloped countries. A country with little or no industrial tradition can hardly produce innovating entrepreneurs. Such entrepreneurs can emerge and work only when a certain level of development is already achieved and people look forward to change and progress. Innovating entrepreneurs played the key role in the rise of modern capitalism through their enterprising spirit, hope of money making, ability to recognize and exploit opportunities, etc.

2. **Adoptive or imitative entrepreneurs:** These are often disparagingly referred to as 'copy cats'. They observe an existing successful system and replicate it in a manner where all the deficiencies of the original business model are addressed and all its efficiencies are retained.

This kind of entrepreneurs are characterised by readiness to adopt successful innovations created by innovative entrepreneurs. These types of entrepreneur are revolutionary entrepreneurs with the different that instead of innovating the changes themselves, they just imitate the technology and techniques innovated by others. These entrepreneurs are most suitable for developing countries because such countries prefer to imitate the technology, knowledge and skill already available in more advanced countries. The Cochin Shipyard is a good example of the result of imitative entrepreneurship. The Shipyard has been constructed using the innovative technology provided by the Mitsubishi Heavy Industries Ltd. of Japan. Imitative entrepreneurs are most suitable for the underdeveloped nations because in
these nations people prefer to imitate the technology, knowledge and skill already available in more advanced countries. In highly backward countries there is shortage of imitative entrepreneurs also. People who can imitate the technologies and products to the particular conditions prevailing in these countries are needed.

Sometimes, there, is a need to adjust and adopt the new technologies to their special conditions. Imitative entrepreneurs help to transform the system with the limited resources available. However, these entrepreneurs face lesser risks and uncertainty than innovative entrepreneurs. While innovative entrepreneurs are creative, imitative entrepreneurs are adoptive.

Imitative entrepreneurs are also revolutionary and important. The importance of these humbler entrepreneurs who exploit possibilities as they present themselves and mostly on a small scale must not be under-estimated. In, the first place, such adaptation requires no mean ability. It often involves what has aptly been called subjective innovation that is the ability to do things which have not been

done before by the particular industrialist, even though, unknown to him, the problem may have been solved in the same way by others.
By western standards, an imitative entrepreneur may be a pedestrian figure, an adopter and imitator rather than a true innovator. He is more an organizer of factors of production than a creator. But in a poor country attempting to industrialise, he is nevertheless a potent change producing figure. He can set in motion the chain reaction' which leads to cumulative progress. This humbler type of entrepreneur is important in under developed countries for another reason. These countries are placing great emphasis in their economic planning on small scale industries and decentralized industrial structure.

3. **Fabian entrepreneurs:** These are entrepreneurs that are very careful in their approaches and cautious in adopting any changes. They are not prone to sudden decisions and try to shy away from any innovations or change that doesn't fit their narrative.

Entrepreneurs of this type are very cautious and skeptical while practicing any change. They have neither the will to introduce new changes nor the desire to adopt new methods innovated by the most enterprising entrepreneurs. Such entrepreneurs are shy and lazy. Their dealings are determined by custom, religion, tradition and past practices. They are not much interested in taking risk and they try to follow the footsteps of their predecessors.

4. **Drone entrepreneurs:** These are entrepreneurs who do not like a change. They are considered as 'old school'. They want to do business in their own traditional or orthodox methods of production and systems. Such people attach pride and tradition to even outdated methods of doing business.

Drone entrepreneurship is characterised by a refusal to adopt and use opportunities to make changes in production methods. Such entrepreneurs may even suffer losses but they do not make changes in production methods. They are laggards as they continue to operate in their traditional way and resist changes. When their product loses marketability and their operations become uneconomical, they are pushed out of the market. They are conventional in the sense that they stick to conventional products and ideas. The traditional industries of Kerala are characterised by drone entrepreneurs. The coir and bamboo industries are still in the hands of laggards who refuse to innovate.

B.) Aruther H. Cole Classification:
Aruther H. Cole classifies entrepreneurs as

Empirical*:* He is an entrepreneur hardly introduces anything revolutionary and follows the principle of rule of thumb.

Rational: The rational entrepreneur is well informed about the general economic conditions and introduces charges which look more revolutionary.

Cognitive*:* Cognitive entrepreneur is well informed, draws upon the advice and services of expert and introduces changes that reflect complete break from the existing scheme of enterprise.

C.) Classification on the Basis of Ownership:

Private*:* Private entrepreneur is motivated by profit and it would not enter those sectors of the economy in which prospects of monetary rewards are not very bright.

Public entrepreneurship: In the underdeveloped countries government will take the initiative to share enterprise.

D.) Classification Based on the Scale of Enterprise:

Scale: The classification is especially popular in the underdeveloped countries. Small entrepreneurs do not possess the necessary talents and resources to initiate large scale production and introduce revolutionary technological changes.

Large Scale*:* In the developed countries most entrepreneurs deal with large scale enterprises. They possess the financial and necessary enterprise to initiate and introduce new technical changes. The result is the developed countries are able to sustain and develop a high level of technical progress.

Other than above mentioned classification, entrepreneurs can also be classified as:

1. Business Entrepreneurs:

Business entrepreneurs are the one who conceive an idea for a new product or service and then create a business to materialise or convert their idea into reality.

2. Trading Entrepreneurs:

They are those entrepreneurs who take up trading activities but are not concerned with its manufacturing i.e. they concentrate on marketing and not production.

3. Industrial Entrepreneurs:

Are those entrepreneurs who identify the potential needs of the customers and manufacture the product or service to meet marketing needs i.e. they concentrate more on Production and not marketing.

4. Corporate Entrepreneurs:

Are those entrepreneurs who demonstrate the innovative skills by converting the Idea into reality organizing and also managing the undertaken business.

5. Agricultural Entrepreneurs:

The entrepreneurs who take up agricultural activities like raising & marketing Crops, fertilizers, insecticides and other related agricultural products and services are called Agricultural entrepreneurs.

6. Technical Entrepreneurs:

These are those entrepreneurs who develop new and improved quality of goods. Because of their craftsmanship and concentrate more on production than marketing.

7. Non-Technical Entrepreneurs:

These are the ones who are not concerned with the technical aspects of the Product. They are more into introducing new styles of advertising, sales promotion etc…,

8. Professional Entrepreneurs: They are entrepreneurs who are interested in establishing a business but do not like to manage or operate the business once it is established. They prefer to sell the set Business and start another venture with the sales proceeds.

9. Pure Entrepreneurs:

The entrepreneurs who undertake business activities for personal satisfaction in terms of status, ego or work and are generally interested in psychological and economic rewards are called as pure entrepreneurs.

10. Induced Entrepreneurs:

These are those entrepreneurs who take up the entrepreneur activities as they are attracted by various policy measures of the government in the form of financial assistance, Infrastructure facilities, Concessions, exemptions and incentives.

11. Motivated Entrepreneurs:

Entrepreneurs with the desire for self fulfilment are motivated entrepreneurs.

12. Spontenous Entrepreneurs:

Individuals who are bold, confident, have strong conviction, have the ability to take the risk and initiate business and moreover start the business as they are naturally talented or have the inborn talent are called as spontaneous entrepreneurs.

13. Growth and Super growth Entrepreneurs:

Entrepreneurs who take up high growth industries are growth entrepreneurs whereas, super growth entrepreneurs show enormous growth of performance in their venture.

14. First Generation Entrepreneurs:

The individuals who are the pioneers or who starts a business unit by putting together their innovative skills and technology and produce marketable product or service are first generation entrepreneurs.

15. Modern Entrepreneurs:

Entrepreneurs who change with the change for demand in the market and undertake those business which modify with the changing are modern entrepreneurs.

16. Classical Entrepreneurs:

They are those entrepreneurs who are concerned with customers and satisfaction of their needs by development and marketing of self-supporting business.

17. Innovative Entrepreneurs:

Individuals who are characterized by information, aggression, analytical Skills have the ability to introduce products with new and innovative features such entrepreneurs are innovative entrepreneurs.

18. Institutional Entrepreneurs:

These are created by government through various financial and promotional institutions for the development of the nations.

19. Forced Entrepreneurs:

These are entrepreneurs who are driven out from other vocations but to satisfy their physiological needs or to run their life are forced to start their business.

20. Inheritance Entrepreneurs:

Businesses which are inherited by family members are inherited ventures and the entrepreneur who accepts the business the way it is run by his forefathers are Inheritance Entrepreneurs.

TYPES OF ENTREPRENEURS

According to the types of business

1. Business entrepreneur
2. Trading entrepreneur
3. Industrial entrepreneur
   - (i) Large
   - (ii) Medium
   - (iii) Small and
   - (iv) Tiny
4. Coroporate entrepreneur
5. Agricultural entrepreneur
   - (i) Placation
   - (ii) Horticulture
   - (iii) Dairy
   - (iv) Forestry
6. Retail entrepreneur
7. Services entrepreneur

According to the use of technologies

1. Technical entrepreneur
2. Non-technical entrepreneur
3. Profession entrepreneur
4. High-tech entrepreneur
5. Low-tech entrepreneur

According to the motivation

1. Pure entrepreneur
2. Induced entrepreneur
3. Motivated entrepreneur
4. Spontaneous entrepreneur

According to the growth

1. Growth entrepreneur
2. Super-growth entrepreneur

According to the stages of development

1. Pure entrepreneur
2. Induced entrepreneur
3. Motivated entrepreneur

According to area

1. Urban entrepreneur
2. Rural entrepreneur

According to gender and age

1. Men entrepreneurs
2. Women entrepreneurs
   - (i) Young entrepreneurs
   - (ii) Old entrepreneurs
   - (iii) Middle-aged entrepreneur

According to the sale of operation

1. Small-scale entrepreneur
2. Large-scale entrepreneur

Other or unclassified

1. Professional entrepreneurs
2. Non-professional entrepreneurs
3. Modern entrepreneurs
4. Traditional entrepreneurs
5. Skilled entrepreneurs
6. Non-skilled entrepreneurs
7. Imitating entrepreneurs
8. Inherited entrepreneurs
9. Forced entrepreneurs
10. National entrepreneurs
11. Internal entrepreneurs
12. Bureaucratic entrepreneurs
13. Entrepreneur entrepreneurs
14. Immigrant entrepreneurs

Functions of Entrepreneurship

(i) Innovation and Creativity – Innovation generally refers to changing processes or creating more effective processes, products and ideas. For businesses, this could mean implementing new ideas, creating dynamic products or improving your existing services. **Creativity** is defined as "the tendency to generate or recognize ideas, alternatives, or possibilities that may be useful in solving problems, communicating with others. Creativity and innovation have always been recognized as a sure path to success. Entrepreneurs think **outside of the box** and explore new areas for cost-effective business solutions.

(ii) Risk taking and Achievement – Entrepreneurship is a process in which the entrepreneur establishes new jobs and firms, new Creative and growing organization which is associated with **risk,** new opportunities and achievement. It results in introducing a new product or service to society. In general, entrepreneurs accept four types of risks namely Financial Risk, Job Risk, Social & Family Risk & Mental & Health Risk, which are as follows:

(a) Financial Risk – Most of entrepreneurs begin by using their own savings and personal effects and if they fail, they have the fear of losing it. They take risk of failure.

(b) Job Risk – Entrepreneurs, not only follow the ideas as working situations, but also consider the current risks of giving up the job & starting a venture. Several entrepreneurs have the history of having a good job, but gave it up, as they thought that they were not cut out for a job.

(c) Social and Family Risk – The beginning of entrepreneurial job needs a high energy which is time consuming. Because of these undertakings, he/she may confront some social and family damages like family and marital problems resulting on account of absence from home and not being able to give adequate time to family.

(d) Mental Health Risk – Perhaps the biggest risk that an entrepreneur takes it is, the risk of mental health. The risk of money, home, spouse, child, and friends could be adjusted but mental tensions, stress, anxiety and the other mental factors have many destructive influences because of the beginning and continuing of entrepreneurial activity. This can even lead to depression, when faced with failure.

(iii) Organization and Management – The entrepreneurial organization is a simple organizational form that includes, one large operational unit, with one or a few individuals in top management. Entrepreneurial management means the skills necessary to successfully develop and manage a business enterprise. A small business start-up under an owner-manager is an example of an **entrepreneurial organization.** Here, the owner-manager generally maintains strict control over business operations. This includes directing the enterprise's core management functions. **According to Mintzberg,** these include the **interpersonal roles, informational roles and decision-making roles**. The smaller the organization, the more concentrated these roles are in the hands of the owner-manager. The entrepreneurial organization is generally unstructured.

(iv) Research – An entrepreneur is a practical dreamer and does a lot of ground-work before taking a leap in his/her ventures. In other words, an entrepreneur finalizes an idea only after considering a variety of options, analysing their strengths and weaknesses by applying analytical techniques, testing their applicability, supplementing them with empirical findings, and then choosing the best alternative. It is then that he/she applies the ideas in practice. The selection of an idea, thus, involves the application of **research methodology**.

(v) Overcoming Resistance to Change – New innovations are generally opposed by people because it makes them change their existing behaviour patterns. An entrepreneur always first tries new ideas at his/her level. It is only after the successful implementation of these ideas that an entrepreneur makes these ideas available to others for their benefit. His/her will power, enthusiasm and energy help him/her in overcoming the society's resistance to change.

(vi) Catalyst of Economic Development – An entrepreneur plays an important role in accelerating the pace of economic development of a country, by discovering new uses of available resources and maximizing their utilization. Today, when India is a fast developing economy, the contribution of entrepreneurs has increased multi-fold.

Barriers in entrepreneurship

Every entrepreneur aims to follow their passionate approach and convert it into a successful business by earning more and more profits and creating goodwill in the market. But many barriers can block the path to the success of entrepreneurship. So there is no other option for an entrepreneur to be successful without solving or removing these major barriers in their respective field.

The list of some such major barriers that may arise while starting or running a new business is given below:

1. Managing Finances

A great barrier that arises while starting a new business is managing the finances that can make things difficult for the entrepreneurs. There are lots of ideas that come to the entrepreneurs' minds but converting those ideas into the business needs enough amount of finance. There should be a stable and regular source of finance to keep the production process smooth. If this first and the most critical barrier can be solved easily then the future barriers can also be tackled.

2. Inadequate Market Experience

One of the most common barriers that are faced by many entrepreneurs is not having adequate knowledge about the respective field of their business. An entrepreneur must have enough experience related to the industry by working in the required sector. So, before starting a new venture, it is required to collect enough knowledge about the market conditions, nature of the business firm, demand and supply of that particular good or service, etc., instead of rushing into the business by seeing the success of others.

Other than this, if an entrepreneur has adequate expertise then it will be helpful for him/her to get success in their career. Although, we know there is always a certain risk involved in a business and the required set of knowledge reduces this risk by creating a strong base.

3. Human Resource Problem

A person can't perform all the tasks by himself/herself even if he/she has enough knowledge. In the same manner, an entrepreneur can't do all the jobs alone to run the organization. A business needs to have a strong base of skilled and knowledgeable human resources or employees. Employees are the most important assets for an organization. They help in the growth of the business. The organizations may have similar fixed and current assets but these are the human assets that make the organization different from its competitors. But for this purpose, the employees must be enough experienced and dedicated to the organization so that top-level productivity and efficiency can be achieved.

An entrepreneur has a large set of applicants who can do a particular job but choosing the best from them is not an easy task. Other than this, the entrepreneur should also see that the candidates who are applying need to have a vision that suits them in the favor of the organization.

When a new venture is started, the management of finance is also a very herculean task because of the limited availability of capital. So, the entrepreneur should never pay a huge salary for the sake of an experienced employee in the beginning.

Because of all these reasons, it is ideal to remove the barrier of human resources so that the business can be run smoothly and the employees can perform their jobs adequately.

4. Non-strategic Planning

Having inadequate planning regarding strategies can also create a great barrier in the path of success of entrepreneurship. There are lots of entrepreneurs who started a business just as a hobby without having a long-term vision and plan. Bad strategies and planning may lead to the failure of the business or can cause a huge loss to the business. It is the prime reason why most entrepreneurs give up in the first year of business operation.

A person should have a solid passion and dedication for being a successful entrepreneur and to create a strategic vision to achieve the goals.

A perfect strategic plan is based on the target customers, market conditions, consumers' tastes, and preferences, proper use of human, financial, and other resources, future goals, etc. All these things are required in the perfect combination for a business to be successful.

5. Lack of Capacity

Many entrepreneurs can't use the opportunities perfectly because of their limited capacity. It can be a major hurdle in the success of entrepreneurship. This hurdle can be arising due to the lack of education, knowledge, willingness, and other important aspects. In simple terms, the absence of zeal and motivation may lead to the failure of most of the new ventures.

Apart from knowing the issues related to market risks and competitions, tons of hard work and passion is also what a business demands to deal with different problems and circumstances that arise in the organization. This is very useful in the growth and development of the business and in increasing the company's turnover and goodwill.

6. Political Barriers

The politics of a country or region can also become a barrier to the success of entrepreneurship. So an entrepreneur should check out the policies of the government and the incentives offered by the government to the entrepreneurs. This is because every government does not provide socio-economic facilities to entrepreneurs. Other than this, it is also required to take the interest in economic development seriously.

Some governments provide special security to the new ventures to protect them from the market competition and so that they can survive in the market. Hence, an entrepreneur must evaluate the political factors carefully before set-up the business firm. It is very helpful in the long-term survival and growth of the business.

7. Lack of Practical knowledge

Many youngsters believe that an entrepreneur must have an adequate educational background to make the business successful, but the reality is completely different. Most of the schools and colleges focus on bookish knowledge instead of providing practical knowledge to the students. Every entrepreneur may have the same bookish knowledge but it is their practical knowledge that makes them unique in this highly competitive market. Lack of this practical knowledge can also become a barrier in the path to the success of a business.

Entrepreneurship is an art that is more based on practical knowledge instead of theoretical. So, every entrepreneur must gain enough practical knowledge of the market conditions and other relevant things before trying their hands in the business.

8. Not Having the Right Team

Business is not the work of an individual. It is teamwork in which every person has different skills and jobs which are based on those skills. So, it can be said that a business is run by the contribution of all the individuals. The right team or teammates or can say employees are the most important aspect for the success of a business. It is one of the biggest barriers for a start-up to find good employees. It is often seen that most of the new ventures work on a small scale which makes it more difficult for them to get experienced, skilled, and good employees. To solve this problem, an entrepreneur should make the required measures so that he/she can assign the jobs to those employees who can boost the business by their effective and efficient performance.

9. Corrupt Business Problems

One of the most common problems with a business especially with new ventures is the corrupt environment. If the environment of the country is corrupt then it does not inspire the young employees and entrepreneurs. In many countries or regions, the unhealthy and non-supportive government creates lots of issues. That's why it is suggested to check out the regulations and stringent compliance before starting a new venture. This corruption of the government can become a major hurdle in the success of a business on a pretty higher level. So, an entrepreneur needs to take things seriously.

10. Difficulty in Reaching to More People

One more barrier that an entrepreneur can face is the difficulty in reaching more people or attracting more customers to inform them about the products or services in which the firm is dealing. A business can lose its path of success if it will not be able to advertise its products and reach the required community in the target market. A successful foundation can only be set up if there will be no barrier between the audience and experts. So, there is the need of having a strong foundation to achieve the planned goals. This problem can be solved by developing effective jobs so that good marketing and advertising strategies can be formed. This will be further helpful in providing information to a large number of people about the organization and its products.

11. Fear of Failure

To get the profit, an entrepreneur has to bear some sort of risk also and it is the entrepreneurs' capacity of bearing the risk which decides the amount of profit that the business can earn. So, it can be said that it is the state of mind which helps an entrepreneur in getting a clear vision so that the risk can be faced without fear. But when a new venture is set up then there is always a fear about the arrangement of initial funds, investors, and many other things. This fear creates pressure on the minds of the entrepreneurs and increases their tension of being failed. The fear of failure can become a huge barrier in starting a new business or running it smoothly. So it's all up to the attitude of the entrepreneurs how they deal with their fear and overcome it to become successful.

12. Few Opportunities

When a new business is started, there are very fewer opportunities available for the entrepreneur. It is also not easy for an entrepreneur to figure out these limited opportunities in the beginning. To get more opportunities, a business must start growing. Limited opportunities can be a hurdle to entrepreneurship. There is almost zero possibility of being successful without having opportunities. Opportunity is also one of the major factors of SWOT analysis.

13. No Training

There is always a great demand in the market for a trained employee. A company always searches for such employees who are trained in their particular job. But for an entrepreneur, it is not enough to have the knowledge or training of a particular field. He/she must be the package of knowledge in all the fields. However, it does not mean that the entrepreneur should be a jack of all trades, master of none because it can also create a barrier in starting entrepreneurship. If the entrepreneurs will not have perfection in a particular field then they will not be able to decide what business they should start. So in this way, it can be said that an entrepreneur should be trained in all works related to the business environment but must have mastery in a particular field also.

14. No Backup Plan

Before starting a new thing, we must have a backup plan in the case of the failure of our main plan otherwise it can create huge trouble for us. In the same manner, when an entrepreneur starts a new venture then he/she must be ready to face the failure, and to deal with them the entrepreneur must have a backup plan or can say plan B. But most entrepreneurs start their business without having a backup plan which can create a barrier for them as they are always at a high risk of failure. Plan B is not only required for the failure of a project but also for the failure of the whole business plan.

15. Limited Technical Skills

Nowadays, most business ideas are based on technology. In such an era, it is easy to start a business for a technical guy but if you have a lack of technical knowledge or skills then it can be a great barrier to you. Limited technical skills can cause a loss to the business. The entrepreneur may hire a particular person to get advice on the issues related to the technical field but it is not always beneficial for the business. Also, it demands additional costs to the organization.

16. Lack of Motivation

The biggest motivation for any business is Money. When an entrepreneur starts a business then it takes some time to make the market value of the business. During this period the profit earned by the business is not so high which may reduce the motivation of the entrepreneur as well as of the employees of the organization.

17. Psychological Pressure

In countries like India, another barrier is psychological pressure in the mind of entrepreneurs. This pressure can be because of any reason such as parental pressure, societal pressure, financial pressure, carrier pressure, etc. One of the biggest examples of these pressures is the pressure created by the family and society to be successful in the carrier. This creates the fear of being failed which gives birth to mistakes and then these mistakes can cause huge losses to the firm.

https://www.easymanagementnotes.com/entrepreneurship-meaning-process-functions-and-barriers/

Entrepreneurs vrs. Manager

Parameters	**Entrepreneur**	**Manager**
Meaning	An entrepreneur is a person who builds a new organisation by gathering data (i.e. land, labour and capital) for manufacturing purposes.	By the term 'manager' we mean a person who gets the things done through his assistants, with the purpose of achieving business goals efficiently.
Focuses on	Setting up a business	Running the daily operations
Status in organisation	Owner	Employee
Kind of benefit applicable	Profit earned from running business	Salary earned from managing daily business operations
Decision making	Intuitive	Calculative
Driving force	Creativity and Innovation	Preserving status quo
Risk orientation	Risk taker	Risk averse

Primary motivation	Achievement	Power

Entrepreneur vs Manager	
Entrepreneur	**Manager**
Entrepreneur is visonary and bears all financial risks.	Manager works for salary, and does not have to bear any risks.
Focuses on starrting and expanding the business ideas	Focus on daily smooth functioning of business
Key motivation for Entrepreneur is achievments	Mangers motivation comes from the power that comes with their position
Reward for all the efforts is profit he earns from the enterprise	Remuneration is the salary he draws from the company
Entrepreneur can be informal and casual	Manager's approach to every problem is very formal

Forms of Business Ownership: Sole proprietorship, partnership forms and others

There are three basic forms of business ownership: sole proprietorship, partnership, and corporation. The form that is chosen can affect the profitability, risk, and value of the firm.

Sole Proprietorship

Sole Proprietorship in simple words is a one-man business organization. It is the type of entity that is fully owned and managed by one natural person (not a legal person/entity) known as the sole proprietor. The business and the man are the same, it does not have a separate legal entity.

A sole proprietorship usually does not have to be incorporated or registered. It is the simplest form of business organizations and the ideal choice to run a small or medium scale business. Let us look at some important features of a proprietorship.

Features of Sole Proprietorship

1] Lack of Legal Formalities

A sole proprietorship does not have a separate law to govern it. So there are not many special rules and regulations to follow. It does not require incorporation or registration of any kind. In most cases, only a license is required to carry out the desired business.

And just like in its formation, there is hardly any legal process involved in its closure. Overall it allows for ease of doing business with minimum hassles.

2] Liability

Since there is no separation between the owner and the business, the liability of the owner is also unlimited. So if the business is unable to meet its own liabilities, it will fall upon the proprietor to pay them. All of his personal assets (like his car, house, other properties etc) may have to be sold to meet the liabilities of the business.

3] Risk and Profit

The owner is the only risk bearer in a sole proprietorship. Since he is the only one financially invested in the company, he must also bear all the risk. If the business fails or suffers losses he will be the one affected.

However, he also enjoys all the profits from the business. He does not have to share his profits with any other stakeholders since there are none. So he must bear the full risk in exchange for enjoying full profits.

4] No Separate Identity

In legal terms, the business and the owner are one and the same. No separate legal identity will be bestowed upon the sole proprietorship. So the owner will be responsible for all the activities and transactions of the business.

5] Continuity

Just as we saw above the business and the owner have one identity. So a sole proprietorship is entirely dependent on its owner. The death, retirement, bankruptcy. insanity, imprisonment etc will have an effect on the sole proprietorship. In most of such cases, the proprietorship will cease to exist and the business will come to an end.

Advantages of Sole Proprietorship

- A proprietor will have complete control of the entire business, this will facilitate quick decisions and freedom to do business according to their wishes
- Law does not require a proprietorship to publish its financial accounts or any other such documents to any members of the public. This allows the business a great deal of confidentiality which is sometimes important in the business world
- The owner derives the maximum incentive from the business. He does not have to share any of his profits. So the work he puts into the business is completely reciprocated in incentives
- Being your own boss is a great sense of satisfaction and achievement. You are answerable only to yourself and it is a great boost to your self-worth as well

Disadvantages of Sole Proprietorship

- One of the biggest limitations of a sole proprietorship is the unlimited liability of the owner. If the business fails it can wipe out the personal wealth of the owner as well and affect his future business prospects too
- Another problem is the limited capital a sole proprietor has access to. His own personal savings and money he can borrow may not be enough to expand the business. Banks and financial institutions are also wary of lendings to proprietorships
- The life cycle of a sole proprietorship is undecided and attached to its owner. If the owner is incapacitated in any way it has a negative effect on the business, and it may even lead to the closure of the business. A sole proprietorship cannot carry on without its proprietor.
- A sole proprietor also has a limited managerial ability. He cannot be an expert in all the fields of the business. And limited resources may mean that he cannot even hire competent people to help him out. This may lead to the business suffering from mismanagement and poor decisions.

Partnership

In India, we have a definite law that covers all aspects and functioning of a partnership, The Indian Partnership Act 1932. The act also defines a partnership as "the relation between two or more persons who have agreed to share the profits from a business carried on by either all of them or any of them on behalf of/acting for all"

So in such a case two or more (maximum numbers will differ according to the business being carried) persons come together as a unit to achieve some common objective. And the profits earned in pursuit of this objective will be shared amongst themselves.

The entity is collectively called a "Partnership Firm" and all the individual members are the "Partners". So let us look at some important features.

Features of a Partnership

1] Formation/Contract

A partnership firm is not a separate legal entity. But according to the act, a firm must be formed via a legal agreement between all the partners. So a contract must be entered into to form a partnership firm.

Its business activity must be lawful, and the motive should be one of profit. So two people forming an alliance to carry out charity and/or social work will not constitute a partnership. Similarly, a partnership contract to carry out illegal work, such as smuggling, is void as well.

2] Unlimited Liability

In a unique feature, all partners have unlimited liability in the business. The partners are all individually and jointly liable for the firm and the payment of all debts. This means that even the personal assets of a partner can be liquidated to meet the debts of the firm.

If the money is recovered from a single partner, he can, in turn, sue the other partners for their share of the debt as per the contract of the partnership.

3] Continuity

A partnership cannot carry out in perpetuity. The death or retirement or bankruptcy or insolvency or insanity of a partner will dissolve the partnership. The remaining partners may continue the partnership if they so choose, but a new contract must be drawn up. Also, the partnership of a father cannot be inherited by his son. If all the other partners agree, he can be added on as a new partner.

4] Number of Members

As we know that there should be a minimum of two members for a partnership. However, the maximum number will vary according to a few conditions. The Partnership Act itself is silent on this issue, but the Companies Act, 2013 provides clarity.

For a banking business, the number of partners must not exceed ten. For a business of any other nature, the maximum number is twenty. If the number of partners increases it will become an illegal entity or association.

5] Mutual Agency

In a partnership, the business must be carried out by all the partners together. Or alternatively, it can be carried out by any of the partners (one or several) acting for all of them or on behalf of all of them. So this means every partner is an agent as well as the principal of the partnership.

He represents the other partners in some cases so he is their agent. But in other circumstances, he is bound by the actions of any of the other partners aking him the principal as well.

Types of Partners

Not all partners of a firm have the same responsibilities and functions. There can be various types of partners in a partnership. Let us study the types of partners and their rights and duties.

- **Active Partner**: As the name suggests he takes active participation in the business of the firm. He contributes to the capital, has a share in the profit and also participates in the daily activities of the firm. His liability in the firm will be unlimited. And he often will act as an agent for the other partners.
- **Dormant Partner**: Also known as a sleeping partner, he will not participate in the daily functioning of the business. But he will still have to make his share of contribution to the capital. In return, he will have a share of the profits. His liability will also be unlimited.
- **Secret Partner**: Here the partner's association with the firm is not public knowledge. He will not represent the firm to outside agents or parties. Other than this his participation with respect to capital, profits, management, and liability will be the same as all the other partners.
- **Nominal Partner**: This partner is only a partner in the name. He allows the firm to use the name of his firm, and the attached goodwill. But he in no way contributes to the capital and hence has no share in the profits. He does not involve himself in the firm's business. But his liability too will be unlimited.
- **Partner by Estoppel**: If a person makes it out to be, through their conduct or behavior, that they are partners in a firm and he does not correct them, then he becomes a partner by estoppel. However, this partner too will have unlimited liability.

Company

A **company** is a legal entity that people form. It allows them, usually as shareholders, to buy and sell goods and services. In most cases, companies want to make a profit. However, some companies are non-profit organizations.

Just like humans, companies can hire and fire workers. They can also buy shares in other companies, sue, and be the target of a lawsuit. Additionally, companies can borrow and lend money, and own property.

As legal entities, companies have similar rights and responsibilities to people. They may be responsible for human rights violations. They may also exercise human rights issues against individuals or the state.

Characteristics of a Company

The company has several distinct characteristics; the significant ones are discussed here:

Separate Legal Entity

A company is a separate legal entity from its members who constitute it. It can hold, purchase and sell properties and enter into contracts in its own name. It is an artificial legal person who can sue aid be sued. Companies are owned by shareholders and they elect the Board of Directors, who run the company. The board, in turn, selects the management. Thus the shareholders exercise only indirect control over the affairs of the company. The separation of ownership from the management some-times results in a conflict of interests between owners and management. The best the shareholders can do is to change some of the directors through a vote in the annual general meeting subsequent to any such conflict.

Limited Liability

The liability of the shareholders of a company is limited to the nominal value of the shares held by them. In the event of liquidation, the maximum loss of a shareholder is equal to the nominal value of the shares held by him. The creditors have no claim on the personal assets of the shareholders in the event of a liquidation.

Transferability of Shares

The shares of a joint-stock company are freely transferable. It does not require any permission from the company or consent of other shareholders. The shares of listed companies can be sold or purchased on the stock exchange and ownership transferred without any difficulty. However, in case of a private limited company, the transfer of shares is subject to the restrictions given in the company's articles.

Ability to Acquire a Broad Capital Base

Following are significant factors that enable a company to raise large amount of capital

1. The nominal value of shares is kept small, as a result of which investment of any size is possible.
2. Limited liability minimizes the risk of the investors and makes investment attractive and safer.

An Artificial Person created by Law

A company is called an artificial person because it does not take birth as a natural person but it comes into existence through the law. The company possesses only those properties which are conferred upon it by its Memorandum of Association (Charter).

Continuous Existence

The companies generally have a continuous existence irrespective of changes in ownership. In the cases of sole proprietorship and partnership, change in ownership means the dissolution of the original business and formation of a new business.

Common Seal

Being an artificial person, a company can act through natural persons only. The acts of a company are authorized by the "common seal". The "common seal" is the official signature of the company. A document not bearing the common seal is not binding on the company.

OTHERS

Corporations (C-Corp and S-Corp)

Corporations are the most complex business structure. A corporation is a legal entity that is separate and independent from the people who own or run the corporation, namely shareholders. A corporation has the ability to enter into contracts separate from that of the shareholders, but it also has certain responsibilities such as the payment of taxes. Corporations are generally more appropriate for larger established companies with multiple employees or when other factors apply (i.e. corporation sells a product or provides a service that could expose the business to sizable liability). Ownership is designated by issuing shares of stock.

The two types of corporations are C-Corps and S-Corps. The major difference among the two types of corporations is the tax treatment of the two entities:

Taxation (C-Corp): For federal income tax purposes, a C-Corp is recognized as a separate taxpaying entity, thus the entity files its own tax return (Form 1120). A c-corporation is subject to corporate income tax on any corporate profits (entity pays taxes). Shareholders pay personal income tax on the corporate profits distributed by the corporation to the owners. As a result, C-corps are subject to "double taxation."

Taxation (S-corp): S-Corps elect to pass corporate income, losses, deductions and credit through to their shareholders for federal tax purposes. However, the entity is required to report income, losses, gains, deductions, credit, etc. on Form 1120S. Shareholders of S corporations report the corporation's income and losses on their personal tax returns pay federal income tax at their individual tax rates. Thus, S-Corps avoid double taxation.

Liability: A corporation is a legal entity that is "immortal," meaning it does not terminate upon the shareholders death. Corporation shareholders have limited liability as they are not personally liable for debts and obligations incurred by the company. Shareholders cannot lose more money than the amount they invested in the corporation. Similar to the provisions of an LLC, shareholders should be careful not to "pierce the corporate veil." Personal checking accounts should not be used for business purposes, and the corporate name should always be used when interacting with customers.

Formation: Corporations are more complex entities to create, have more legal and accounting requirements and are more complex to operate than sole proprietorships, partnerships, or LLCs. One of the major disadvantages of a corporation is the high level of governance and oversight by the board of directors. Often times, this prolongs the decision making when multiple shareholders or investors are involved.

Pros of Corporations:

• Corporate shareholders have limited liability, meaning the entity is responsible for all liabilities the company incurs.
• Usually a favorable formation for investors.

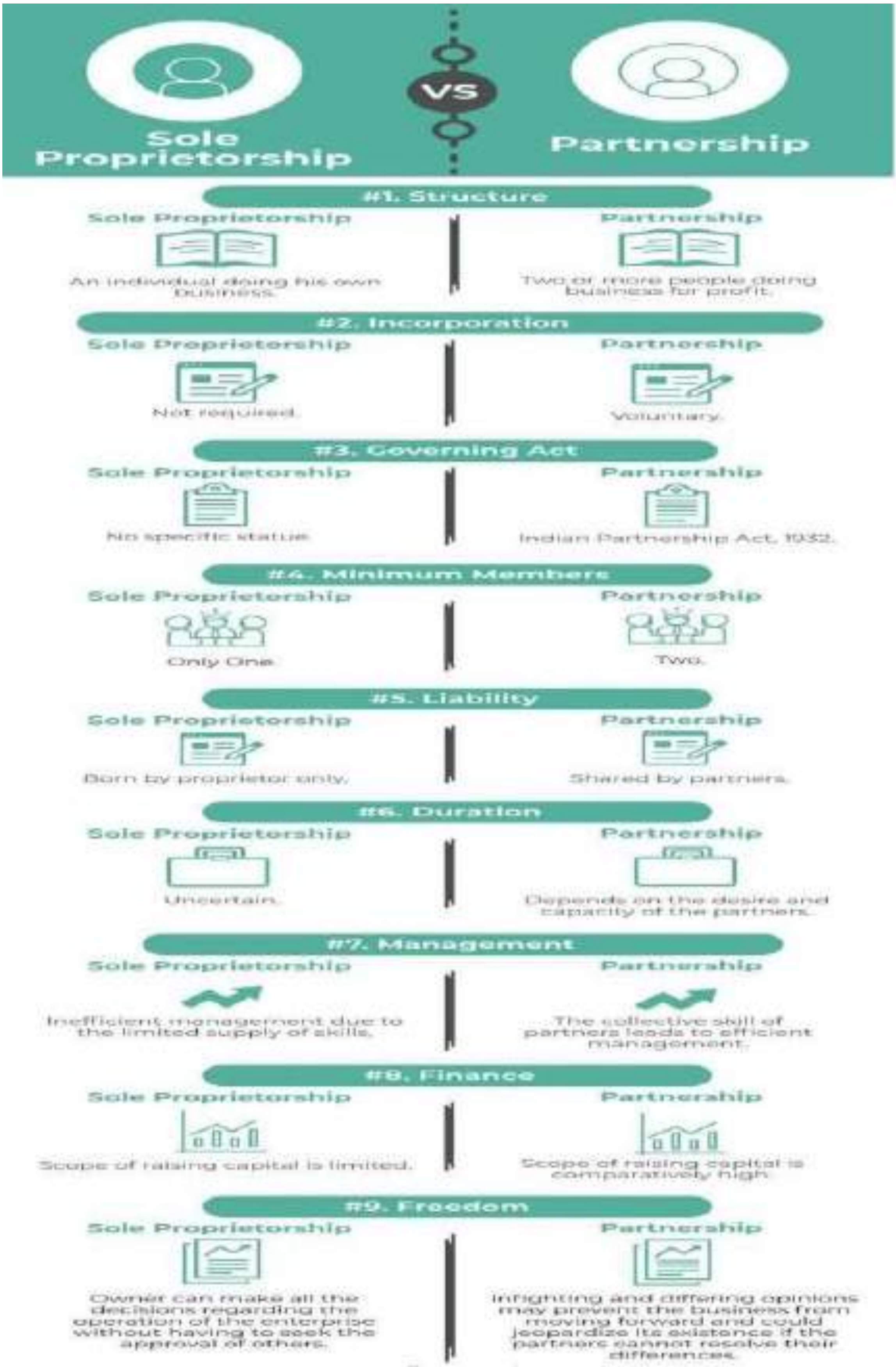
Cons of Corporations:

• The process to establish the business is more rigorous and costly.
• Earnings are subject to "double taxation", meaning that earnings are taxed at the entity level and the individual level upon distribution to shareholders.
• High level of governance and oversight by the board of directors.

KEY POINT

1. A sole proprietorship is a business owned by only one person.
   - Advantages include: complete control for the owner, easy and inexpensive to form, and owner gets to keep all of the profits.
   - Disadvantages include: unlimited liability for the owner, complete responsibility for talent and financing, and business dissolves if the owner dies.

2. A general partnership is a business owned jointly by two or more people.
   - Advantages include: more resources and talents come with an increase in partners, and the business can continue even after the death of a partner.
   - Disadvantages include: partnership disputes, unlimited liability, and shared profits.



Types of Industries

https://www.flexiprep.com/NIOS-Notes/Secondary/Business-Studies/NIOS-Business-Studies-Ch-2-Industry-and-Commerce-Part-2.html

Concept and Types of industry

Industry means place where the resources are converted into finished or useful products. There are basically three types of industries.

I) Primary industry

II) Secondary industry

III) Tertiary industry

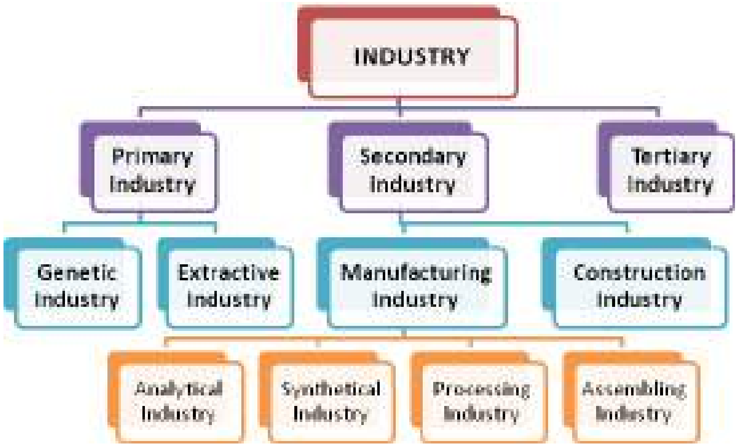
I) **Primary industry:** This type of industry involves in the activities related to extraction and production of natural resources. It also performs activities relating to the reproduction and development of living organisms. It is of two types:

1. Genetic industry: This industry involves breeding of plants and animals for their use in reproduction process like poultry farming, fish hatching.
2. Extractive industry: This industry performs the activities relating to the extraction of products from natural resources like farming, mining etc.

II) **Secondary industry:** This industry converts the goods which have been extracted or produced at the primary industry into final products. It is further classified into two parts:

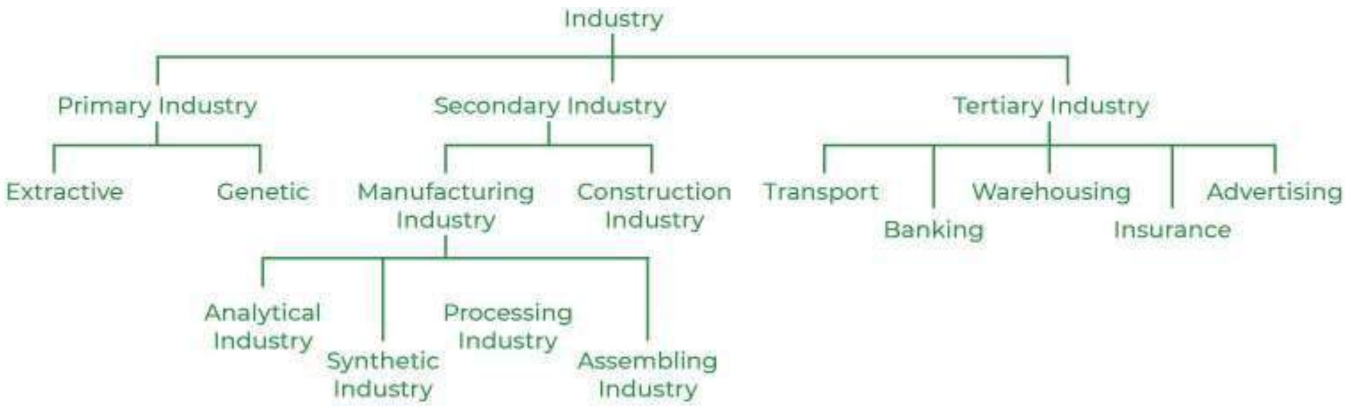
1. Manufacturing industry: This industry generally makes products for final consumption and helps in creating form utility. It covers four types of industry.
2. Analytical industry: This type of industry involves in the separation of different elements from the same product. Like in oil refining.
3. Synthetical industry: This industry involves mixing various products and changing into a new product.
4. Processing industry: This industry deals in processing of product in further stages like making cheese or butter from milk.
5. Assembling industry: This industry creates a new product by assembling different parts together like I case of car assembling.

III) **Tertiary industry:** This type of industry provides support service to the primary as well as secondary industry in trading process.



In other way we can categorised as-

The industry is concerned with the processing and production of goods and services with the motive of earning profit. In simple words, it involves changing the raw materials into finished products for the purpose of selling them to the customers. The goods sold by industry can be either used by other companies for further production or can be used by the end consumers for final consumption. The goods produced in the former case are known as producer goods. For example, machinery, tools, etc. However, the goods produced in the latter case are known as consumer goods. For example, bread, milk, butter, cloth, groceries, medicines, etc. An organization can also produce materials that can be used by other companies to further process and convert them into finished goods. These goods are known as intermediate goods. For example, rubber, copper, plastic, etc. Industry can be further classified into three parts: **primary, secondary, and tertiary industry**.



Primary Industry

The primary industry involves the extraction of raw materials or natural resources from the earth and the reproduction of living organisms. The economic operations of a business under the primary sector usually depend on the nature of a specific place where the extraction or reproduction is being done. Therefore, we can say that the economic operations of a primary industry revolve around the natural resources available on earth. The primary industries produce or create products that are sold to the public. For example, farming, mining, fishing, forestry, agriculture, crop production, animal husbandry, etc. The primary sector of less advanced economies and includes activities or businesses from these areas. As a country develops, it becomes less dependent on primary industry and more on secondary and tertiary. Some of the workers in the primary industry are hunters, farmers, coal miners, etc.

Primary industry is further divided into two categories:

1. **Extractive Industry:** As the name suggests, the extractive industry involves the extraction of products from natural resources and manufacturing of finite raw materials that industry cannot replenish through cultivation. In emerging countries like India, the primary industry is usually the most important sector. For example, agriculture is the most crucial part of India's primary

industry, and animal farming is important in Africa. Other essential extractive industries are mining, fishing, hunting, lumbering, and farming.

2. **Genetic Industry:** The genetic industry involves rearing and breeding of living species like birds, plants, etc., developing raw materials, and then improving them through human involvement in the production process. Genetic industry examples involve seeds and nursery companies, poultry farms, fish hatcheries, breeding farms, livestock management and forestry.

Secondary Industry

The secondary industry uses the raw materials extracted in the primary sector and then converts them into the finished product. Therefore, the secondary industry consists of construction and manufacturing industries. The products manufactured under secondary industries are either consumed by the end customer or used as raw material by other industries for further processing or production. For example, getting wood from forests is a primary industry; however, making furniture from wood is a secondary industry. The businesses under secondary industry usually use massive machinery in their production plants for manufacturing goods and even use human resources for packaging and distribution of goods to retailers, wholesalers, etc., at different locations.

Secondary industry is further divided into two categories: Manufacturing and Construction Industry.

Manufacturing Industry

These are the industries that are involved in the process of converting raw materials or semi -finished products into finished products. Manufacturing industries create form utility as it changes the form of raw materials into finished goods. For example, cotton is a raw material that is converted in clothes under the manufacturing industry, iron is converted into benches, sugarcane into sugar, wheat int o bread, etc.

The two types of goods produced in the manufacturing industry are:

- **Consumer Goods:** Consumer goods are goods that are consumed directly by the end consumer and are used for day-to-day consumption. Examples of consumer goods are clothes, hand wash, soap, oil, ghee, bread, butter, etc.
- **Industrial Goods:** Industrial goods are the goods that are built or produced for manufacturing consumer goods. For example, machines, equipment, etc., are used by industries to manufacture consumer goods such as bread, butter, soap, oil, etc.

The manufacturing industry is further divided into four categories:

1. **Analytical Industry:** The industry in which a basic raw material is broken down into several parts for manufacturing multiple products is known as analytical industry. For example, different products like gasoline, kerosene oil, diesel, petroleum, etc., are formed by breaking crude oil into different parts and processing them separately.
2. **Synthetic Industry:** The industry in which manufacturers mix two or more materials for manufacturing a new product is known as the synthetic industry. For example, products like paint, paper, soap, cement, sanitizers, etc., are manufactured by mixing two or more than two materials.
3. **Processing Industry:** The industry in which manufacturers process a raw material through different production stages and then manufacture the finished good is known as the processing industry. For example, the Sugar industry processes sugarcane through different stages to manufacture sugar for consumers and other industries.
4. **Assembling Industry:** The industry in which the organizations take different finished products and combine them to form a new finished product is known as the assembling industry. For example, computer companies buy different finished products like CPU, motherboard, software, etc., from different companies and produce a new product.

Construction Industry

These are the industries that are involved with the construction of roads, dams, buildings, etc., for the development of an economy. The construction industry use products of the manufacturing industry, such as steel, iron, cement, etc. The products of construction industries are unique in a way that their products cannot be moved or transferred from one place to another. For example, a dam can be built at a fixed place and cannot be moved as per the choice of anyone.

Tertiary Industry

The tertiary industry involves providing services and facilitating a smooth flow of goods and services in the market. The tertiary industry is also known as the service industry, and it helps the primary and secondary industries with their activities. It means that the tertiary industry provides services to different primary and secondary industries to support their activities.

Different types of services provided by the service industry are:

1. **Transport:** Transportation services help the primary and secondary industries by facilitating the movement of goods from one place to other. Different modes of transport used by tertiary industry are air, water, land, rail, etc.
2. **Banking:** Through banking, the tertiary industry provides credit facilities and finance to different trading firms and industries for business expansion, survival and growth.
3. **Warehousing:** Warehousing services of the tertiary industry provides the primary and secondary industry with storage facilities. It means that the primary and secondary industries can store their produced goods until they are distributed.
4. **Insurance:** Insurance services of the tertiary industry involve providing coverage to the primary and secondary industries with different types of risks while running the business.
5. **Advertising:** Advertising services help the primary and secondary industries by providing information to the customers about the company or business and its goods and services.

https://www.lendingkart.com/blog/types-of-industries/

Concept of Start-ups Entrepreneurial support agencies at National, State, District Level(Sources): DIC,NSIC,OSIC, SIDBI, NABARD, Commercial Banks, KVIC etc.

- A start-up is a company that's in the initial stages of business.
- Founders normally finance their start-ups and may attempt to attract outside investment before they get off the ground.
- Funding sources include family and friends, venture capitalists, crowd funding, and loans.
- Start-ups must also consider where they'll do business and their legal structure.
- Start-ups come with high risk as failure is very possible but they can also be very unique places to work with great benefits, a focus on innovation, and great opportunities to learn.

The support agencies has set up various centres or institutes to impart training and development to entrepreneurs, so as to improve their knowledge, attitudes, and skills.

For the development of entrepreneur a number of specialized agencies have been set up by the state and central governments which are as follows:
Government also plays a very important role in developing entrepreneurship. Government develop industries in rural and backward areas by giving various facilities with the objective of balances regional development. The government set programmes to help entrepreneurs in the field of technique, finance, market and entrepreneurial development so that they help to accelerate and adopt the changes in industrial development. Various institutions were set up by the central and state governments in order to fulfil this objective.

The Government has set up various centres or institutes to impart training and development to entrepreneurs, so as to improve their knowledge, attitudes, and skills.

1. District Industry Centres (DICs):

Governments, both Central and State, have in the past taken a number of measures for the development of small and village industries, but the actual achievement has been far below expectation, because the focus of attention for industrial development was mainly on large cities and state capitals.

Also the multiplicity of institutions involved in the development of small industries and complicated systems and procedures made the job of promoting the industrial units an uphill task for small entrepreneurs.

Hence, it was felt necessary to establish a development agency which could provide all services and facilities to village and small industries under one roof. Accordingly, the DICs were established on May 8, 1978 in order to cater the needs of small units.

Each district has a DIC at its headquarter. The main responsibility of DIC is to act as the chief coordinator or multi-functional agency in respect of various Government departments and other agencies. The prospective small entrepreneur would get all assistance from DIC for setting up and running an industry. The metropolitan cities of Delhi, Mumbai, Kolkata and Chennai have been kept outside the purview of the DIC.

Organisational Setup of DICs:

Usually each DIC consists of:

i. One General Manager of the rank of Joint Director of Industries as the head.

ii. Four Functional Managers, of whom, three would be in the areas of economic investigation, credit and village industries. The fourth functional manager would be entrusted with the responsibility in any of the areas like raw materials/marketing/training etc. depending on the specific requirement of each district.

iii. Three Project Managers to provide technical services in the area relevant to needs of the district concerned. Their role is to facilitate modernisation and up- gradation of technology in the small sector.

At the sub-division level, there could be Assistant Director of Industries and Industry Promotion Officer, but these positions do not exist in every state.

Functions of DICs:

The role of DIC is mainly promotional and developmental. To attain this purpose, it has to provide needed services and support to small and village industries.

i. Identification of Entrepreneurs – DIC develops new entrepreneurs by conducting entrepreneurial motivation programmes throughout the district especially in small towns.

ii. Selection of Projects – DIC offers technical advice to new entrepreneurs for the selection of projects suitable to them.

iii. Registration under SSI – DIC provides provisional and permanent registration to the new entrepreneurs. After the selection of projects, entrepreneurs are issued with provisional SSI Registration which is essential for obtaining assistance from the financial institutions. Once a unit comes into being, then the unit is registered permanently.

Provisional registration is awarded for two years in the first instance and thereafter, can be renewed every year taking into consideration the fact that the party is taking effective steps to install the unit, but this renewal is limited to only two times.

iv. Clearances from Various Departments – It takes the initiative to get clearances for the project from various departments and takes follow up measures to get speedy power connection.

v. Recommending Applications to Various Other Departments – The District Industry Centre recommends the applications of SSI units for availing loans and working capital facilities to commercial banks and other financial institutions.

The DIC also recommends the application of SSI units to National Small Industries Corporation for the purchase of machinery on hire-purchase/lease basis, after assessing the technical and economic viability of the industrial units.

vi. Arranging Seed Money – The DIC, through various associated corporations, provides seed money to entrepreneurs who are technically qualified for setting up their small industrial ventures, but are not in a position to muster their own capital as a part of their contribution towards the financial assistance which they are to get from the banks or from the financial institutions.

vii. Assistance to Raw Material Supplies – DIC makes necessary recommendations to the concerned raw material suppliers and issues the required certificates for the import of raw materials, wherever necessary. This facility enables the SSI units to get raw material on controlled rates and as per their assessed requirements.

viii. Financial Assistance – Financial assistance to SSI units is being provided through various schemes of the Industries Department. Under Prime Minister's Rozger Yojana for Unemployed Youth, eligible unemployed youths get financial assistance in the form of loan and subsidy on easy terms and conditions up to Rs.95,000 for industrial, business and service activities.

ix. Subsidy Schemes – DIC assists SSI units and rural artisans to get subsidies such as power subsidy, interest subsidy, subsidy under IRDP etc. from various institutions.

DIC also provides subsidy to such SSI units which purchase testing machinery/equipment for carrying out different tests. This helps the SSI units to keep a constant watch on their products for ensuring quality.

x. Grant of Purchase Preference – The DIC allows 15% purchase preference to SSI units and 5 percent to large and medium units in respect of purchase of products by various departments approved for availing the concession.

xi. Assistance to Village Artisans and Handicrafts – DIC arranges for financial assistance with the lead bank or nationalised banks of the respective areas for village artisans.

xii. Entrepreneurship Development Programmes /Seminars – Entrepreneurship Development Programmes have been an important aspect of industrialisation since the entrepreneur may be highly qualified, but he still may not be familiar with the different aspects and steps required in setting up an industrial venture.

District Industry Centres have been conducting these programmes in association with Small Industries Service Institutes and various Technical Consultancy Organisations to develop entrepreneurial skills in the young entrepreneurs.

xiii. Import and Export Assistance – The items covered under the restricted and limited permissible categories can be imported against the Import Licence. The Import Licences are issued by the concerned Import and Export authorities on the recommendations of the General Manager of the concerned DIC.

The SSI units desirous of exporting goods are required to get registered with the concerned Export Promotion Council for availing various facilities. The application for this registration is routed through the General Manager of the concerned DIC.

xiv. Fairs and Exhibitions – With a view to give wide publicity to the industrial products being manufactured by the various industrial centres throughout India, the Govt., of India organises India International Trade Fair every year. This helps the industrial units to export their products to the outer world, identify prospective buyers and open new marketing avenues.

2. National Small Industries Corporation (NSIC):

The National Small Industries Corporation (NSIC) was set up in 1955 with the objective of supplying machinery and equipment to small enterprises on a hire- purchase basis and assisting them in procuring Government orders for various items of stores. The supply of machines on hire-purchase is in a way an offer of funds, an offer of foreign exchange facilities, guidance on adopting modernised technology for improved methods of production and combination of all.

NSIC takes upon itself the entire purchase procedure, starting from locating competent suppliers to delivery of machines. In case of imported machines, NSIC obtains clearance from Director General to arrange foreign exchange, obtain import licence, opens the letter of credit and looks after the customs requirement and clearance of machines.

The Corporation's Head office is at Delhi and it has four regional offices at Delhi, Mumbai, Chennai and Kolkata and eleven branch offices. It has one central liaison office at Delhi and depots and sub-centres.

Functions of NSIC:

The NSIC has taken up the challenging task of promoting and developing small scale industries almost from scratch and has adopted an integrated approach to achieve the socio-economic objectives.

The followings are the main functions of NSIC:

i. To develop small scale units as ancillary units to large-scale industries.

ii. To provide machines to small scale industries on hire-purchase basis.

iii. To assist small enterprises to participate in stores purchase programme of the Central Government.

iv. To assist small industries with marketing facilities.

v. To distribute basic raw materials through their depots.

vi. To import and distribute components and parts to actual small scale users in specific industries.

vii. To construct industrial estates and establish and run prototype production cum-training centres.

3. OSIC Odisha Small Industries Corporation Ltd (OSIC).

OSIC is the nodal agency for supply of raw-material to MSMEs in the state for taking up infrastructure development of the MSME sector.

The development of a separate cell is likely to strip off the rights vested now with the state-run Odisha Industrial Infrastructure Development Corporation (IDCO). It may be noted that IDCO, which is also the nodal agency of government for providing land to large projects, was established in 1981 with the specific objective of creating infrastructure facilities in identified industrial estate and areas.

The prime function of IDCOis establishment of industrial areas, industrial estates, functional industrial estates, growth centres at different strategic locations and to provide infrastructure.

Prior to formation of IDCO in 1981 for creation of industrial infrastructural facility, the portfolio of development of industrial estates was with the OSIC.

4. Small Industries Development Bank of India (SIDBI):

With a view of ensuring larger flow of financial and non-financial assistance to the small-scale sector, the Government of India set up the Small Industries Development Bank of India (SIDBI) under a special act of the Parliament in October 1989 as a wholly owned subsidiary of IDBI.

The bank commenced its operations from April 2, 1990 with its head office in Lucknow. The SIDBI has taken the outstanding portfolio of the IDBI relating to the small-scale sector worth over Rs. 4000 crores. The authorised capital of SIDBI is Rs. 250 crores with a provision to increase it to Rs. 1000 crores.

Functions of SIDBI:

The SIDBI was set up to function as the principal financial institution for the promotion, development and financing of industry in the small scale sector and for co-ordinating the functions of institutions engaged in similar activities. It has taken over the responsibility for administering Small Industries Development Fund and National Equity Fund which were earlier administered by the IDBI.

The SIDBI has outlined the following three areas as immediate thrust areas:

i. To initiate steps for technological upgradation and modernisation of existing units.

ii. To expand the channels for marketing the products of SSI sector in domestic and international markets.

iii. To promote employment oriented industries especially in semi-urban areas to create more employment opportunities and thereby checking migration of people to urban areas.

The SIDBI has outlined the following activities as its functions:

i. Refinancing of loans and advances extended by primary lending institutions such as State Financial Corporations (SFCs), State Industrial Development Corporations (SIDCs) or State Industrial and Investment Corporations (SIICs) and banks.

ii. Discounting and rediscounting of bills.

iii. Extension of seed capital or soft loan assistance under National Equity Fund, Seed Capital under Mahila Udayam Nidhi Scheme.

iv. Granting direct assistance and refinance for financing exports of SSI sector.

v. Extending financial support to State Small Industries Development Corporation (SSIDC).

vi. Extending financial support to National Small Industries Corporation (NSIC).

5. Commercial Banks and Entrepreneurial Development:

Commercial banks play an important role in the growth and development of economy in general and the enterprise sector, in particular. The role of banks in promotion and development of business and industry has become all the more important in post liberalisation era.

These days, commercial banks have not confined themselves to mere extension of finance to small entrepreneurs, but have shown genuine concern for their progress and development. They have now entered the challenging field of promoting new small scale entrepreneurs through Entrepreneurship Development Programmes.

In their new role as promoters of small scale sector, they have accepted yet another challenging task. They are now holding EDPs in collaboration with specialised institutions such as DIC, SISI, TCO etc. with a view to identify entrepreneurs, especially in backward areas and train and monitor them to start new ventures.

Commercial Banks-

- Organised under the Banking Companies Act, 1956
- They operate on a commercial basis and its main objective is profit.
- They have a unified structure and are owned by the government, state, or any private entity.
- They tend to all sectors ranging from rural to urban
- These banks do not charge concessional interest rates unless instructed by the RBI
- Public deposits are the main source of funds for these banks

The commercial banks can be further divided into three categories:

1. **Public sector Banks** – A bank where the majority stakes are owned by the Government or the central bank of the country.
2. **Private sector Banks** – A bank where the majority stakes are owned by a private organization or an individual or a group of people
3. **Foreign Banks** – The banks with their headquarters in foreign countries and branches in our country, fall under this type of bank

Function of Commercial Bank:

The functions of commercial banks are classified into two main divisions.

(a) Primary functions

Accepts deposit : The bank takes deposits in the form of saving, current, and fixed deposits. The surplus balances collected from the firm and individuals are lent to the temporary requirements of the commercial transactions.

Provides loan and advances : Another critical function of this bank is to offer loans and advances to the entrepreneurs and business people, and collect interest. For every bank, it is the primary source of making profits. In this process, a bank retains a small number of deposits as a reserve and offers (lends) the remaining amount to the borrowers in demand loans, overdraft, cash credit, short-run loans, and more such banks.

Credit cash: When a customer is provided with credit or loan, they are not provided with liquid cash. First, a bank account is opened for the customer and then the money is transferred to the account. This process allows the bank to create money.

(b) Secondary functions

Discounting bills of exchange: It is a written agreement acknowledging the amount of money to be paid against the goods purchased at a given point of time in the future. The amount can also be cleared before the quoted time through a discounting method of a commercial bank.

Overdraft facility: It is an advance given to a customer by keeping the current account to overdraw up to the given limit.

Purchasing and selling of the securities: The bank offers you with the facility of selling and buying the securities.

Locker facilities: A bank provides locker facilities to the customers to keep their valuables or documents safely. The banks charge a minimum of an annual fee for this service.

Paying and gathering the credit : It uses different instruments like a promissory note, cheques, and bill of exchange.

6. National Bank for Agriculture and Rural Development (NABARD):

The National Bank for Agriculture and Rural Development (NABARD) was established in 1982 for providing credit for the promotion of agriculture, small-scale industries, cottage and village industries, handicrafts and other rural crafts and other allied economic activities in rural areas with a view to promote integrated rural development and secure prosperity of rural areas.

On its establishment, NABARD has taken over the entire undertaking of the Agricultural Refinance and Development Corporation and has taken over from the Reserve Bank its refinancing functions in relation to the State Co-operative Rural Banks. This Bank is now the co-ordinating agency in relation to the Central Government, Planning Commission, State Governments and institutions engaged in the development of small-scale industries, village and cottage industries, rural crafts etc. for giving effect to the various policies and programmes relating to rural credit.

The capital of the NABARD is Rs. 500 crores, subscribed by the Central Government and the Reserve Bank in equal proportions.

Management of NABARD:

In terms of the Act, the Board of Directors will consist of 15 members to be appointed by the Central Government in consultation with the Reserve Bank. Besides the chairman and the managing director, three directors from the Central Board of the Reserve Bank, three officials of the Central Government, two officials of the State Governments and five directors from experts in rural economics, rural development, handicrafts and village and cottage industries etc. and persons with experience in the working of co-operative banks and commercial banks.

Functions of NABARD:

i. NABARD provides short-term refinance assistance for periods not exceeding 18 months to state co-operative banks, regional rural banks and other financial institutions for a wide range of purposes including marketing and trading relating to rural economy.

These short term loans can be converted by the NABARD into medium term loans for periods not exceeding seven years under conditions of drought, famine or other natural calamities, military operations or enemy action.

ii. NABARD can grant medium-terms loans to the State Co-operative Banks and Regional Rural Banks for periods extending from 18 months to 7 years for agriculture and rural development.

iii. NABARD is empowered to provide by way of refinance assistance long term loans extending up to maximum period of 25 years to the State Land Development Bank, Regional Rural Banks, Scheduled Commercial Banks, State Co-operative Banks or any other financial institution approved by the Reserve Bank for giving loans to artisans, small-scale industries, village and cottage industries etc.

7. Khadi and Village Industries Commission (KVIC):

Khadi and Village Industries Commission was established in 1953 with the primary objective of developing khadi and village industries and improving rural employment opportunities. Its wide range of activities include training of artisans, extension of assistance for procurement of raw materials, marketing of finished products and arrangement for manufacturing and distribution of improved tools, equipments and machinery to producers on concessional terms.

KVIC provides assistance to Khadi and Village Industries which require low capital investment and ideally suited for manufacturing utility goods by using locally available resources. There are many specified village industries such as processing of cereals and pulses, leather, matches, gur and khandsari, non-edible oils and soaps, bee-keeping, village pottery, carpentry and blacksmithy etc.

KVIC's policies and programmes are executed through State Khadi and Village Industries Boards registered under the Societies Registration Act, 1960 and Industrial Cooperative Societies registered under State Cooperative Societies Act. Activities involving pioneering type of work such as developing new industries in hilly, backward and inaccessible areas are undertaken by KVIC directly.

For the development of entrepreneur a number of specialized agencies have been set up by the state and central goverenments which are as follows:

- Small Industries Service Institutes (SISI)
- Small Industries Development Organisations (SIDO)
- National Small Industries Corporation
- Small Industries Extension Training Institute.
- Entrepreneurship Development Institute of india
- Institute for Rural Management and Adminstration
- National Institute for Entrepreneurship and Small Business Development (NIESBUD)
- National Alliance of young entrepreneurs (NAYA)

Technology Business Incubators (TBI) and Science and Technology Entrepreneur

Incubators provide resources and services to entrepreneurs, including working space and offices, technical expertise, management mentoring, assistance in compiling an effective business plan, shared administrative services, technical support, business networking, and advice on intellectual property, sources of financing, markets, and strict admission/exit rules.

A technology business incubator (TBI) is an organisational setup that nurtures technology. based and knowledge driven companies by helping them survive during the startup period in. the company's history, which lasts around the initial two to three years.

A technology business incubator (or TBI) is a type of business incubator focused on organizations that help startup companies and individual entrepreneurs which use modern technologies as the primary means of innovation to develop their businesses by providing a range of services, including training, brokering and financing. In several countries, including India, China, and the Philippines there have been government initiatives to support TBIs.[2][3] Organizations under the title of technology business incubator often receive funding or other forms of support from the national government.

EXAMPLE-

1. AMITY INNOVATION INCUBATOR
2. THE INNOVATION CONTINUUM
3. INDIAN ANGEL NETWORK

Science & Technology Entrepreneurship Park (STEP)

The Science Parks and similar initiatives help in creating an atmosphere for innovation and entrepreneurship; for active interaction between academic institutions and industries for sharing ideas, knowledge, experience and facilities for the development of new technologies and their rapid transfer to the end user.

The Science & Technology Entrepreneurs Park (STEP) programme was initiated to provide a re-orientation in the approach to innovation and entrepreneurship involving education, training, research, finance, management and the government. A STEP creates the necessary climate for innovation, information exchange, sharing of experience and facilities and opening new avenues for students, teachers, researchers and industrial managers to grow in a trans-disciplinary culture, each understanding and depending on the other's inputs for starting a successful economic venture. STEPs are hardware intensive with emphasis on common facilities, services and relevant equipments.

The major objectives of STEP are to forge linkages among academic and R&D institutions on one hand and the industry on the other and also promote innovative enterprise through S&T persons.

Objectives

- To forge a close linkage between universities, academic and R&D institutions on one hand and industry on the other.
- To promote entrepreneurship among Science and Technology persons, many of whom were otherwise seeking jobs soon after their graduation.
- To provide R&D support to the small-scale industry mostly through interaction with research institutions.
- To promote innovation based enterprises.

Facilities and Services Provided by STEPs

- It offers facilities such as nursery sheds, testing and calibration facilities, precision tool room/central workshop, prototype development, business facilitation, computing, data bank, library and documentation, communication , seminar hall/conference room , common facilities such as phone, telex, fax, photocopying. It offers services like testing and calibration, consultancy.

- Training, technical support services, business facilitation services, database and documentation
- services, quality assurance services and common utility services .

UNIT-2
MARKET SURVEY AND OPPORTUNITY IDENTIFICATION

Business Planning

- SSI, Ancillary Units, Tiny Units, Service sector Units
- Time schedule Plan, Agencies to be contacted for Project Implementation
- Assessment of Demand and supply and Potential areas of Growth
- Identifying Business Opportunity
  Final Product selection

Business Planning

- A business plan is a document that defines in detail a company's objectives and how it plans to achieve its goals. A business plan lays out a written road map for the firm from marketing, financial, and operational standpoints. Both startups and established companies use business plans.
- A business plan is a documented strategy for a business that highlights its goals and its plans for achieving them. It outlines a company's go-to-market plan, financial projections, market research, business purpose, and mission statement. Key staff who are responsible for achieving the goals may also be included in the business plan along with a timeline.
- The purpose of a business plan is three-fold: It summarizes the organization’s strategy in order to execute it long term, secures financing from investors, and helps forecast future business demands.

Purposes of a Business Plan

1. Securing financing from investors.

Since its contents revolve around how businesses succeed, break-even, and turn a profit, a business plan is used as a tool for sourcing capital. This document is an entrepreneur's way of showing potential investors or lenders how their capital will be put to work and how it will help the business thrive.

All banks, investors, and venture capital firms will want to see a business plan before handing over their money, and investors typically expect a 10% ROI or more from the capital they invest in a business.

Therefore, these investors need to know if – and when – they'll be making their money back (and then some). Additionally, they'll want to read about the process and strategy for how the business will reach those financial goals, which is where the context provided by sales, marketing, and operations plans come into play.

2. Documenting a company's strategy and goals.

A business plan should leave no stone unturned.

Business plans can span dozens or even hundreds of pages, affording their drafters the opportunity to explain what a business' goals are and how the business will achieve them.

To show potential investors that they've addressed every question and thought through every possible scenario, entrepreneurs should thoroughly explain their marketing, sales, and operations strategies – from acquiring a physical location for the business to explaining a tactical approach for marketing penetration.

To show potential investors that they've addressed every question and thought through every possible scenario, entrepreneurs should thoroughly explain their marketing, sales, and operations strategies –

from acquiring a physical location for the business to explaining a tactical approach for marketing penetration.

These explanations should ultimately lead to a business' break-even point supported by a sales forecast and financial projections, with the business plan writer being able to speak to the why behind anything outlined in the plan.

3. Legitimizing a business idea.

Everyone's got a great idea for a company – until they put pen to paper and realize that it's not exactly feasible.

A business plan is an aspiring entrepreneur's way to prove that a business idea is actually worth pursuing.

As entrepreneurs document their go-to-market process, capital needs, and expected return on investment, entrepreneurs likely come across a few hiccups that will make them second guess their strategies and metrics – and that's exactly what the business plan is for.

It ensures an entrepreneur's ducks are in a row before bringing their business idea to the world and reassures the readers that whoever wrote the plan is serious about the idea, having put hours into thinking of the business idea, fleshing out growth tactics, and calculating financial projections.

4. Getting an A in your business class.

Speaking from personal experience, there's a chance you're here to get business plan ideas for your Business 101 class project.

What does a business plan need to include?

1. Business Plan Subtitle
2. Executive Summary
3. Company Description
4. The Business Opportunity
5. Competitive Analysis
6. Target Market
7. Marketing Plan
8. Financial Summary
9. Team

SSI

Small Scale Industries (SSI) are those industries in which the manufacturing, production and rendering of services are done on a small or micro scale. These industries make a one-time investment in machinery, plant, and equipment, but it does not exceed Rs.10 crore and annual turnover does not exceed Rs.50 crore.

On 9 May 2007, subsequent to the amendment of the Government of India (Allocation of Business) Rules, 1961, the Ministry of Small Scale Industries and the Ministry of Agro and Rural Industries were merged to form the Ministry of Micro, Small and Medium Enterprises. Thus, the SSIs are included under the Ministry of MSME.

Characteristics of SSI

Ownership

SSI’s generally are under single ownership. So it can either be a sole proprietorship or sometimes a partnership.

Management

Generally, both the management and the control is with the owner/owners. Hence the owner is actively involved in the day-to-day activities of the business.

Labor Intensive

SSI’s dependence on technology is pretty limited. Hence they tend to use labour and manpower for their production activities.

Flexibility

SSI’s are more adaptable to their changing business environment. So in case of amendments or unexpected developments, they are flexible enough to adapt and carry on, unlike large industries.

Limited Reach

Small scale industries have a restricted zone of operations. Hence, they can meet their local and regional demand.

Objectives of SSI

The objectives of the small scale industries are:

- To create more employment opportunities.
- To help develop the rural and less developed regions of the economy.

- To reduce regional imbalances.
- To ensure optimum utilisation of unexploited resources of the country.
- To improve the standard of living of people.
- To ensure equal distribution of income and wealth.
- To solve the unemployment problem.
- To attain self-reliance.
- To adopt the latest technology aimed at producing better quality products at lower costs.

Types of SSI

SSI are primarily categorised into 3 types, based on the nature of work carried out, which are as follows:

Manufacturing Industries

The manufacturing industries manufacture finished goods for consumption or used further in processing. Some examples of such SSIs are food processing units, power looms, engineering units, etc.

Ancillary Industries

Ancillary industries manufacture components for other manufacturers. These manufacturers then assemble the final product. Big companies manufacture finished goods, but they do not generally make all the parts themselves. The vendors of such big companies are ancillary industries.

Service Industries

Service-based industries are not involved in any kind of manufacturing products. They provide services such as repair, maintenance and upkeep of the products after-sales.

Other types of SSIs are as follows:

Export Units

An SSI is considered as an export unit when the exporting is more than 50% of its production.

Cottage Units

The cottage units are considered as SSIs when they do not involve a dedicated facility and are carried out within living spaces or houses of the owners.

Village Industries

An SSI is considered village industries when they are established in rural areas and are not part of the organised sector. Typically, these industries solely depend on human labour for production.

Examples and Ideas of SSIs in India

- Bakeries
- School stationeries
- Water bottles
- Leather belt
- Small toys
- Paper Bags
- Photography
- Beauty parlours
- Spinning and weaving industry
- Coconut oil making
- Cashew nut processing
- Clay products
- Agarbatti making
- Chalk making
- Biodiesel production

- Rice mill
- Toys making
- Honey processing
- Slippers making
- Detergent powder making
- Spices making
- Chocolate making
- Water bottles manufacturing
- Toothpick making
- Xerox and printing
- Pickle manufacturing industry
- Incense stick manufacturing industry
- Paper plate manufacturing industry
- Candle manufacturing

Ancillary Units

An ancillary small industrial unit is the one who supplies not less than 50% of its production to another industry. The term ancillary industry refers to that industry which supports the main industrial activity in a given region or city. Ancillary industries are referred to as those industries which invest in plant and machinery with a capital not exceeding rupees 1 crore.

Ancillary industries do the work of making parts and components used by heavy engineering industries to shape their final product for ultimate consumption. In this way the heavy industries get freed from the pressure of making every part and component and in this way the capital is also saved by them so that they can make investment somewhere else.

Some characteristics of ancillary industries are as follows:
- They supply a minimum 50 % of their production to heavy industries also known as parent industries. The maximum capital they can invest is Rs. 1 Crore.
- The various segments which use ancillary units are for ex: - textile machinery, agriculture machinery, railways, automobile industry, shipbuilding industry, aircraft industry etc.

The advantages that heavy industries have due to ancillary industries are as follows:
> Innovation: Due to the existence of ancillary units the heavy industries save much time for doing innovation in their products rather than simply making smaller units.
> Specialization: When parts or components are made by ancillary units, they are specialized in that product but if this part is to be made in the parent company it will be only its sideline operation.
> No storage needed: The storage problem is solved as it is in the hands of heavy industries to order

only when they require.
> Suppliers can be changed easily: Heavy industries can easily change the suppliers (ancillary industries or units) of components or small parts if quality of product decreases which is not easy in case they are made on their own. Also, if there is a need to change their requirement, they can easily change their suppliers.
> Capital diversification: Heavy industries can use capital saved from while taking off small parts from ancillary industries to research work in their product and thus make more profit as that product will not be available in scarcity as compared to smaller units.

Ancillary Unit:

1. It is that type in which its supplies its 50 % production to the parent unit.
2. Its investment limit is almost one crore.
3. There is assistance required from parents by providing technical and financial support.
4. It can do its business by itself.
5. It is of small scale.

Tiny Units

1. It is that type in which machinery and investment of plant is not more than 25 lac.
2. Its investment limit is almost 25 lac.
3. There is no assistance required.
4. It cannot do its business by itself.
5. It is of large scale.

Difference between Ancillary Units and Tiny Units

Basis of difference	Ancillary industrial units	Tiny units
Definition	Industrial units that have to supply a minimum of 50% of their production to their parent industries are termed ancillary industrial units.	Industries that have a maximum investment of Rs. 25 lakh in their plant and machinery are regarded as tiny industrial units.
Investment limit	The maximum level of investment is Rs. 1 crore.	The maximum level of investment is Rs. 25 lakh.
Obligation	Such units have to supply at least 50% of their production to their parent industries.	No such obligations.
Examples	Industries engaged in the production of machine parts, tools and other intermediate products.	Business units such as small shops, boutiques, STD (subscriber trunk dialling) booths and photocopy centres.

Service sector Units

The service sector is the sector of the economy that produces and offers services. According to the tri-sector macroeconomic theory, there are three major economic sectors – primary, secondary, and tertiary.

1. Primary Sector

The primary sector is composed of industries engaged in the business of gathering raw materials. It includes mining companies, lumber companies, and oil drilling companies – along with the agricultural and fishing industries.

2. Secondary Sector

The secondary sector includes all businesses involved in producing and selling goods – such as auto manufacturers, furniture stores, and clothing retailers.

3. Tertiary Sector

The tertiary sector is the service sector. It includes industries such as the financial services industry, internet technology (IT), and the healthcare and entertainment industries.

The service sector or tertiary sector refers to one of the portions forming the three-sector model of the economic sector. The businesses in the service industry produce intangible goods in the form of service as output delivering to other businesses or consumers.

- The service sector refers to the industry producing intangible goods, the services as output. In many countries, it is the largest and fastest-growing sector.
- It can be categorized into consumer, business, and public services.
- Examples include hospitality, education, information technology, media, and entertainment.
- The tertiary sector dominates the Indian economy. Liberalization and a series of economic changes in the 1990s fuelled the rise of the service industry, transforming an agriculture-dependent economy into one dominated by the service industry.

Types of service sector

There are different types of tertiary sectors and classifications. Moreover, their proportion varies with countries. One of the classifications is organizing them into consumer, business, and public services.

- **Consumer Services**: **Target market** will be the individual consumers. Examples include hospitality and retail.
- **Business Services**: Target market will be the organizations or other businesses. Examples include information technology and marketing services.
- **Public Services**: Deliver services to the general public. Examples include healthcare and education.

Top 10 service sector companies in India:

1	Reliance Industries
2	HDFC Bank

3	ICICI Bank
4	HDFC
5	Tata Consultancy Services
6	Larsen & Toubro
7	State Bank of India
8	NTPC
9	Kotak Mahindra Bank
10	Axis Bank

Time schedule Plan

A **schedule plan**, also called a **schedule management plan**, is an essential component of project management. Project managers create schedule plans to determine the start and end dates of a project, the milestones along the way, and a timeline for completing individual tasks.

Schedule planning can benefit a project in several ways:

- Keep team members abreast of deadlines and responsibilities.
- Stay mindful of time-related costs.
- Monitor how resources are allocated.
- Develop a solid **project plan** that guides the entire team through initiating, executing, and concluding a project.

How to create a schedule plan

Follow the steps below to start building your schedule plan.

1. Define the project's activities and their duration.

In project management, an activity refers to a single stage in the process of bringing a project to completion. Each activity has its own start and end date. Altogether, a project's activities lead to the final deliverable.

Work with your project team and, if possible, subject-matter experts to create a list of all the activities required to complete the project. Account for the individual tasks for each activity and the time required to complete them. Then, you'll need to order the activities in the most logical sequence so that the process can be tracked and managed.

2. Determine the project's resources.

For this step, refer to your sequence of activities to determine what resources you'll need to carry out those activities. Resources might include people, equipment and supplies, funding, information, or facilities. Take into account when resources will become available, and adjust the sequence of activities accordingly.

3. Select a scheduling methodology.

In project management, there are different approaches to scheduling a project. You'll want to select the methodology that makes the most sense for your project by considering all factors that may affect the project timeline.

- **The Critical Path Method** is sequential and linear. Each task must be completed before beginning the task that follows. Industries that use the Critical Path Method include construction, aerospace and defense, and product development.

- **Agile** is an iterative method by which projects are divided into shorter-term development cycles, allowing team members to assess progress at the end of each cycle and make adjustments for the cycle that follows. Industries that use Agile include marketing and advertising, banking and finance, and health care.
- **The Program Evaluation and Review Technique (PERT)** involves identifying the time it takes to complete each task or activity, so that project managers can better estimate the project's completion date. PERT works well for research and development projects, as well as projects that don't involve repetitive activities.

4. Select a software.

Project scheduling software can help teams complete projects in a timely fashion by:
- Providing a central source of reference
- Displaying a visual representation of the overall workflow, task deadlines, how the project is progressing, etc.
- Organizing information about the project
- Automating tasks like sending notifications to individual team members and generating data reports

Collaborate with your team to select the best scheduling software for your project.

5. Establish the schedule baseline.

A schedule baseline is the schedule that project stakeholders approve once all tasks, timelines, and resources have been accounted for. Having a schedule baseline provides the team with a tool for gauging whether the project stays on schedule.

When a task or activity takes longer than is expected, project managers and teams can refer to the schedule baseline to determine how best to get a project back on track.

6. Establish performance measures.

Just as you use the schedule baseline to measure whether a project is progressing on time, you'll need to establish other areas of project performance that you want to measure along the way. These can include:
- Whether the team is reaching the established project milestones
- The quality of the work that team members are completing at each phase of the project
- Team morale and the attitudes they have toward the project
- Resource management

Decide how and when you will measure performance, such as holding meetings with team members and stakeholders to discuss the project or collecting their feedback through a questionnaire.

7. Create contingency plans.

With most projects, some delays and challenges can impede progress. For example, resources crucial to a particular activity may become unavailable, thus resulting in a delay or even an added cost.

Your schedule plan should document potential risks as well as contingency plans, or Plan B, for how the team will handle instances when potential risks become a reality.

8. Identify schedule change control procedures.

In this section of your schedule plan, you will document instances when specific schedule elements can be changed and by whom. You should also include details about the impact of specific schedule changes on the project, including the time to completion, the need for additional resources, or changes to actual tasks and activities. The procedure itself might include several steps, such as:

- Completing a change request
- Assessing the work involved in implementing the change
- Choosing a course of action: denying, accepting, or modifying the change request

9. Establish the format and frequency of progress reporting.

In this section, you will document how team members should submit updates on the project. Include details such as:

- Who will submit progress reports
- How often they should submit reports
- Required information for the reports
- The format for these reports (visual representations or written descriptions)
- Where the reports are submitted

Agencies to be contacted for Project Implementation

In order to be eligible to apply for project, an entity should be:-

i. registered under Indian Trust Acts or any State Society Registration Act or any State Cooperative Societies or Multi-State CO-operative Acts oF the Companies Act 2013 or the Limited Liability partnership Act, 2008.
ii. Government or a semi-Government organization at the State and National Level.

Project Implementing Agency means the responsible agency designated to implement, monitor, and evaluate each city demonstration project under Part 2 of the Project and the term "Project Implementing Agencies" shall refer collectively to every Project Implementing Agency.

All the payments under the contract for the package for which this invitation for Bids is issued shall be made by the Employer {who is also named as **Project Implementing Agency** (PIA) by Ministry of Power/GoI} named in Biding Documents.

Project Implementing Agency (PIA) shall formulate a detailed comprehensive Quality Assurance (QA) plan for the works to be carried out under IPDS scheme with an objective to create quality infrastructure works.

(PIA) shall formulate a detailed comprehensive Quality Assurance (QA) plan for the works to be carried out under DDUGJY scheme with an objective to create quality infrastructure works. It is the responsibility of Watershed Committee to form SHGs in their respective villages under the guidance of Watershed Development Team and Project Implementing Agency.

Project Implementing Agency (PIA): Society for Elimination of Rural Poverty, Department of Rural Development, Government of Telangana 2.

Assessment of Demand and supply and Potential areas of Growth

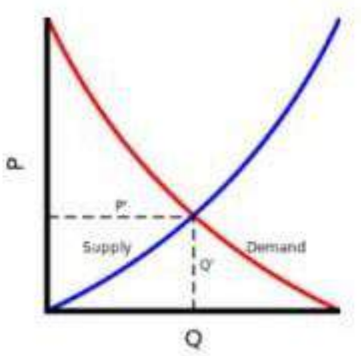
- Demand and supply analysis is the study of price changes with respect to the number of buyers and sellers for a particular product.
- In a state of market equilibrium, the demand curve and the supply curve intersect. At this point, the demand and supply for a commodity are equal.
- The fair price of farm yield is also based on demand-supply. In a perfectly competitive market, farmers are price takers and demand-supply is the price maker. Therefore, governments set a minimum price to protect farmers.

- Supply and demand is an economic model which states that the price at which a good is sold is determined by the good's supply, and its demand. When the supply of a good is equal to its demand (known as economic equilibrium), it reaches a stable price which buyers and sellers can agree on.

- If the supply of a good is higher than its demand, then the price will drop (various sellers will have to compete with each other by offering lower prices, which will in turn create more demand), until eventually the supply and demand equalize. When the supply for a good is greater than the demand, it is referred to as a surplus.

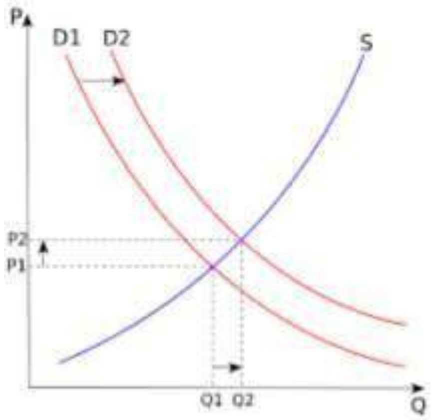
- If the demand for a good is greater than its supply, then the opposite will occur. Suppliers will increase their prices to earn more profit with the products they already have, until eventually the supply and demand reach equilibrium at some peak price. When the demand for a good is greater than the supply, it is referred to as a shortage.

- The law of demand states that (with a few exceptions) as price rises, the quantity demanded of any good or service would be lower.
- The law of supply implies that higher the price received by a supplier, the quantity supplied will rise. Thus, demand is often a downward sloping curve in the price-quantity plane, while supply is an upward sloping curve.



- **What is the Equilibrium Price?**
- The intersection of the supply and demand curve denotes the market equilibrium, which in turn determines the equilibrium levels of price and quantity of the particular good (or service) in the economy.
- If the present demand for a good (or service) in the economy is higher than the equilibrium quantity, the situation is described as that of an excess demand. Excess supply is also defined in a similar fashion.

- **What Causes Shifts in Supply or Demand?**
- Changes in Supply and demand (and thus the equilibrium price and quantity) of any good or service could be governed by a lot of factors, such as: changes in policies, unpredictable shocks to the economy, business cycle fluctuations like a recession or a boom, or even simply over time (long run versus short run). It also depends on the nature of the market (whether the market is perfectly competitive or monopolistic etc.).
- The analysis of all the above could be termed as the study of the supply and demand, or simply, 'Demand Supply Analysis'.



Identifying Business Opportunity

To be successful entrepreneurs, we need to be continually innovating and looking for opportunities to grow our startups. An entrepreneur takes the initiative to start the business activities, by exploring and analyzing or Identification of business opportunities. He also bears the related risks. The real entrepreneur is not risk centered, rather he is opportunity centered. The entrepreneurs explore economic and business opportunities and remain ready to take advantage of those opportunities.

Following are the objectives of the Identification of business opportunities:

1. Identification of opportunities by an entrepreneur, in the context of probable industries and to decide his own role, the scope of work, and relationships, in accordance with the opportunities.
2. To keep watch over the possible market of the commodity or service to be produced.
3. To decide a high-level group of managers, so that entrepreneurial ventures may be started.
4. To make an assessment of financial resources by making financial forecasts, in the context of the process if industrial development.
5. To explore the opportunities for possible entry in other areas.
6. To assess the requirements of labor, capital, and materials for the industries.
7. To find out the possibilities of short term and long term development in various areas of the economy.
8. To have the desire for technical knowledge, awareness towards new opportunities, and acceptance of the changes.
9. To see the possibilities of diverting the available resources towards achieving the **business objectives**.
10. To identify those industries, which are not based on local sources, but which may be economically considered, in view of future requirements.

Spotting market opportunities is essential for business growth and survival. Opportunities are factors that contribute to your growing success. These are typically external and include things like:

- economic trends

- market trends
- shifting or expanding customer base
- changes in government or industry regulations
- changes in partnerships or relationships with suppliers, competitors, etc
- new or changing funding prospects (eg increase in grant funding)

One of the best ways of identifying new business opportunities is to complete a SWOT analysis.

SWOT analysis to identify business opportunities

A SWOT analysis is one of the most popular strategic analysis tools. It focuses on four key factors affecting your business:

- strengths
- weaknesses
- opportunities
- threats facing your business

A SWOT analysis typically provides you with valuable information about your business. It can help you to identify new opportunities and prospects that you can potentially exploit.

For example, if SWOT analysis shows that your opportunities lie in changing market trends, you may want to respond by introducing new products, altering prices of existing goods, adapting your advertising or adding value to the customer experience.

Other ways to detect new business opportunities

Other than using SWOT analysis, to detect new business prospects you can also:

- Segment your customers - divide your customer base into smaller groups that share common characteristics, such as age, gender, location, or lifestyle, attitude or purchasing habits. This allows you to analyse the demand for your products or services and target each segment with specific marketing, promotions and offers, ultimately helping you to seize new business opportunities.
- Analyse your purchase situation - look at distribution channels, payment methods and other factors around purchasing to understand how your customers buy your products or services. Then look at how you can improve them or offer alternatives to bring new customers to your business.
- Analyse your competition - research existing businesses in the market to understand their value proposition and their competitive advantage. Establish your unique selling point to set yourself apart from your competitors. Identify what draws your customers in and promote this.
- Look at other markets - exporting into other countries may open up new opportunities. Consider carefully what the market is like. To increase the likelihood of success, research the competitors, local habits and the likely demand for your products or services.

Consider exploiting other types of business opportunities too. For example, you can:

- franchise your business
- buy an existing business
- find a distributor to increase your market share
- explore licensing, supply chain deals, etc.

Final Product selection

Goods and services are necessary for satisfying the needs of society. Such goods/services or goods along with services or simply an idea can be called as a product. In simple terms, product is anything

that is offered for sale into the marketplace. This makes product an important constituent of marketing and an important consideration in managerial decision-making.
There are various sources for collecting the vast array of information regarding product selection. These sources include government publications. development agencies, management and technical journals, public sector units like ITI, HAL, BHEL, etc, and various research laboratories. In many cases, entrepreneurs may hire specialized institutions or individuals to carry out the process of product selection on their behalf. Similarly, the task of market survey may be conducted by the entrepreneur himself or by hiring experts or consultants.

Criteria to Select a Product

An entrepreneur makes use of several criteria before selecting a product. These are as follows :

1) Technical Know-how :
Technical know-how is an important guiding factor for an entrepreneur while selecting a product. An entrepreneur can easily decide what product should be manufactured if he belongs to the product-related field. Similarly, knowledge in manufacturing or marketing field enables the entrepreneur in selection of an appropriate product.

2) Availability of Market :
The availability of a large market for a particular product also helps in its selection. If the demand for a product is huge, then the market risk of launching it becomes less. Thus, it is very essential that the entrepreneur have a good knowledge about the product's market in terms of how and where the product can be sold.

3) Financial Strength :
Relative financial of the entrepreneur also serves as guiding force in selection of a product. Manufacturing a product generally requires heavy investment in research and development, capacity creation, plant and machinery, etc., which is generally beyond the expenditure capacity of a small-sized firm. Thus, it is advisable for an entrepreneur to analyse its financial strength prior to product selection.

4) Competitive Rivalry :
The returns from a product are greatly influenced by the degree of competition prevailing in the market. Factors like market dominance by the competitors, availability of substitutes, any barrier to entry, etc., play significant role in determining the viability of the product.

5) Product Category :
In many cases, certain products fall in the priority sector category while some others may be reserved for small scale. sectors. The level and extent of competition for these product categories among small sectors is lesser than that seen in other sectors of the economy. There are also certain products which the Government has earmarked as exclusively to be purchased from the small scale sector. In the case of such products, the entrepreneur will definitely give a greater importance to a product which falls in this category.

6) Consistency in Demand :
When there are not many fluctuations in the demand of a product then the market for such products can be considered to be stable. Seasonal products are contrary in the sense that their demand fluctuates a lot. The seasonality of a product also plays a large part in its selection or non-selection as this is directly linked to the stability in the demand for the same. The entrepreneur should definitely prefer a product which has a consistent and stable demand.

7) Restriction on Imports :
The foreign trade policy of the government may restrict the import of some products. In such cases, those products. gain attraction from general public and consequently an entrepreneur should prefer those products that are part of such restricted category.

8) Availability of Raw Materials :

Availability of raw material is a very important factor for selecting a particular product. An entrepreneur should ensure that supplies required for smooth conduct of business operations are mostly available in desired quantity around the year. Moreover, the source of procurement of raw material is also important. Where supplies are to be procured from external sources, the entrepreneur will have to maintain a sufficient quantity of inventory compared to local sources.

9) Government Incentives and Subsidies :

Government often provide a number of subsidies and incentives for the promotion of certain businesses. These incentives and subsidies are generally in the form of tax holidays, exemptions from customs, concessions, etc. An entrepreneur must consider the availability of such governmental schemes as they greatly support an entrepreneur in setting up a new business.

10) Ancillary Products :

When the product is in the nature of an ancillary product (a product required for manufacturing another product), then its increases the attraction for the entrepreneur. This is because the product will have a ready market in the parent industry. For example, an ancillary unit of Maruti.

11) Location of Business :

Business location is also important for selecting a particular product as certain products are earmarked for production in special zones like free trade zones, export promotion zones, etc. Government also provides incentives and tax breaks for such products. Moreover, the location of big consumer markets near the production centers also increases the attractiveness of certain products. Such products will also be selected by the entrepreneur because of the locational advantage.

12) Licensing System :

There are overtime changes in the governmental licensing policies. For some products, it is mandatory for the entrepreneur to have the required license issued by the concerned authority. Under particular conditions, capital addition is also monitored. Moreover, the process of obtaining license is burdensome for certain products. Thus, products which require a lot of licensing approvals will not be very attractive to entrepreneurs.

13) Government Policy :

The selection of product also has to be done keeping in mind the government policies and their likely impact. The entrepreneur should choose a product which falls in a sector with favorable Governmental policies. For example, products which are not socially beneficial like tobacco and alcohol do not receive government support.

UNIT-3
PROJECT REPORT PREPARATION

- Preliminary project report
- Detailed project report, Techno economic Feasibility
- Project Viability

In consideration to specific requirements, a project report must entail relevant information in several formats and present it in an easy to understand and properly structured manner. Prepared in the form of pertinent ranking or reports, or economic trend predictions, project report is the most significant tool for decision-making.

Standard Project Report Format

The measurement of a standard project report should be in A4 size. It should be bounded by elastic cover preferably of white color art paper and should be prepared according to the below mentioned format of a project report:

1. **Title Page**
   - The title page and cover page must display the topic of the project report

2. **Authentic Certificate**
   - The authentic certificate must posses the controller's signature, followed by his name and designation along with full address, institution's name and his department's name.

3. **Summary**
   - The abstract must be a one page short summary of the project report. It should provide an insight of the comprehensive report.

4. **Content Tables**
   - The content table should catalog all the topics which are included in the project report. Among the content items, the authentic certificate and title page should not be included.

5. **Catalog of symbols, acronyms and categorizations**
   - The catalog should include all the standard symbols, acronyms and categorizations used in the project report.

6. **Chapters:**
   This can be classified in three broad categories namely preliminary chapter, chapters introducing the focal theme of the project report and conclusion. The chief content is again classified various lessons with each lesson being further classifies into various segments and sub-segments.

   Make sure that every lesson is given suitable heading. The placing of facts and figures and other statistics in each chapter should be at the instant vicinity of the suggestion where they are quoted. The annotation s should be placed carefully. They should be positioned below the text that they refer to.

7. **Appendices**
   - Appendices are included in a project report to provide additional information which is incorporated in the primary content and serve as a diversion and cloud the fundamental topic. They should be figured using Arabic numerals such as Appendix 2, Appendix 15, Appendix 21, etc. Tables and suggestions featuring in Appendices should be figured and submitted at suitable positions.

8. **References Catalog**

- The References Catalog should be placed below a broad "References" title. The suggestion text should be organized in the alphabetical sequence of the primary writer. The writer's name should be followed by the year of edition and other relevant details.

Formulation of project Report

A project report is an operating document. So that what information and how much information it contains depends upon the size of the enterprise, as well as nature of production.

Project formulation divides the process of project development into eight distinct and sequential stages as below:

- General information
- Project description
- Market potential
- Capital costs and sources of finance
- Assessment of working capital requirements
- Other financial aspects
- Economical and social variables
- Project implementation

The nature of information to be collected and furnished under each of these stages has been given below.

1. General Information

The information of general nature given in the project report includes the following :

Bio-data of promoter : Name and address, qualifications, experience and other capabilities of the entrepreneur. Similar information of each partner if any.

Industry profile : A reference analysis of industry to which the project belongs, e.g., past experience; present status, its organization, its problem etc.

Constitution and organization : The constitution and organization structure of the enterprises; in case of partnership form its registration with registration of firms, certificate from the directorate of industries.

Product Details : Product utility, product range, product design, advantage to be offered by the product over its substitutes if any.

2. Project description

A brief description of the project covering the following aspects should be made in the project report.

Site : Location of the unit; owned,rented or leasehold land; industrial area; no objection certificate from municipal authorities if the enterprise location falls in the residential area.

Raw Material : Requirement of raw material, whether inland or imported, sources of raw material supply.

Skilled Labour : Availability of skilled labour in the area i.e., arrangements for training laborers in various skills.

Utilities : These include –

- Power
- Fuel
- Water
- Pollution Control
- Communication and transportation facility
- Production Process

- Machinery and Equipment
- Capacity of the Plant
- Technology Selected
- Other Common Facilities
- Research and Development

3. Market Potential

While preparing a project report, the following aspects relating to market potential of the product of the product should be stated in the report.

- Demand and Supply position
- Expected Price
- Marketing Strategy
- After Sales Service

Depending upon the nature of the product, provisions made for after-sales should normally in the project report.

4. Capital costs and sources of finance

An estimate of the various components of capital items like land and buildings, plant and machinery, installation costs, preliminary expenses, margin of working capital should be given in the project report.

5. Assessment of working capital requirements

The requirement for working capital and its sources of supply should clearly be mentioned. It is preferred to prepare working capital requirements in the prescribed formats designed by limits of requirements.

It will reduce the objections from Banker's side.

6. Other financial aspects

To adjudge the profitability of the project to be set up, a projected profit and loss amount indicating likely sales revenue, cost of production, allied cost and profit should be prepared.

A projected balance sheet and cash flow statement should also be prepared to indicate the financial position and requirements at various stages of the project.

Breakdown level indicates the gestation period and the likely moratorium required for repayment of the loans. Break-Even point is calculated as

Break-Even point (BEP) = $F/S - V$

Where

F = Fixed Cost

S = Selling Price/Unit

V = Variable Cost/Unit

The break-even point indicates at what even of output the enterprise will break even.

7. Economical and social variables

Every enterprise has social responsibility. In view of the social responsibility of business, the abatement costs, i.e., the costs for controlling the environmental damage should be stated in the project.
In addition the following socio-economic benefits should also be stated in the report.

- Employment Generation

- Import Substitution
- Ancillaration
- Exports
- Local Resource Utilization
- Development of the Area

8. Project implementation

Every entrepreneur should draw an implementation scheme or a time-table for his project to the timely completion of all activities involved in setting up an enterprise. If there is delay in implementation project cost overrun.

Techno-Economic Feasibility Study

Techno-Economic Feasibility Study or "TEFS" means a comprehensive study including the results of resources/reserves estimation and in case of viability, proposed geo-technical, geo- hydrological, mining, and processing/metallurgical methods and financial analysis taking into account technical, engineering, legal, operating, environmental and economic factors as well as

Objective of Techno-Economic Feasibility Studies/ Detailed Project Reports is to determine the technical feasibility and financial viability of the project, assess the risks associated with the project and enumerate imminent actions that are required to be taken. It helps a client get a detailed evaluation of a project.

Coverage:

Techno-Economic Feasibility Studies/ Detailed Project Reports cover the following based on the clients requirement:

- **Markets:** It covers estimated future sales revenue of the project based on estimated sales volumes and price.
- **Raw Materials & Fuel:** It estimates the adequacy of the quality and quantity of the raw materials and fuel for the project, make an estimate of its cost.
- **Plant Siting, Location & Infrastructure:** It assesses the existing infrastructure and actions required to develop the infrastructure necessary to set up the project.
- **Project Technical Concept:** This is core deliverable of the project and covers plant capacity, equipment sizing, storages, plant auxiliaries, system engineering, electrical engineering, civil engineering, Control & Automation engineering, Quality Control & Assurance, Captive Power Plant and Waste Heat recovery system (WHR) based on the project requirement.
- **Logistics:** Inbound and outbound logistics and logistics planning.
- **Environment:** Applicable Regulatory Framework and Environmental Impact of Project.
- **Implementation Planning**.
- **Human Resources:** Requirement and cost.
- **Investment Cost:** Fund requirement.
- **Operating Cost:** It covers cost of raw material, utilities, overheads, etc.
- **Financial Appraisal:** It covers project profitability, IRR, NPV, payback, etc.
- **Risks & Mitigation**.

Methodology:

Broad methodology of execution involves:

- Data collection
- Data analysis
- Report and Project formulation

Benefits:

It helps a client make an investment decision based on:

- Technical feasibility of the project.
- Financial viability of the project.
- The risks associated with the project.
- Actions required for risk mitigation.

UNIT-4
MANAGEMENT PRINCIPLES

- Definitions of management
- Principles of management
- Functions of management (planning, organising, staffing, directing and controlling etc.)
- Level of Management in an Organisation

Definitions of management

According to Follet management is '**the art of getting things done through people**'. Management is a distinct process consisting of activities such as planning, organizing leading and controlling to achieve the desired results. The goals of the enterprise are fulfilled through the use of resources like men, money, materials and machines. Managerial Skills are required to exploit these resources and earn profit, face competition and ensure long survival and growth of an enterprise. Throughout the management function, these objectives must be kept in mind .

The first weakness is that the definition states that management is an art. Art deals with application of knowledge. But management is not merely application of knowledge. It also involves acquisition of knowledge i.e., Science. Managing using intuition or thumb rule is not correct management. The second weakness of this definition is that it does not explain the various functions of management.
George R Terry who defines management as "a process consisting of planning, organizing, actuating and controlling performed to determine and accomplish the objectives by the use of people and resources". According to him, management is a process-a systematic way of doing thing using four managerial functions namely planning, organizing, actuating and controlling. 'Planning' means thinking of the manager's action in advance. The actions of the managers are based on logic, plan or some method rather than hunch. 'Organizing' means coordinating machines, materials and human resources of the organization. 'Actuating' means motivating, directing the subordinates. 'Controlling' means that manager must ensure that there is no deviations from plans.

1.1 CHARACTERISTICS OF MANAGEMENT

The critical analysis of the above definitions, the following characteristics of management evolve.

(1) **Management is a continuous process**: The process of management consists of planning, organizing, directing and controlling the resources to ensure that resources are used to the best advantages of the organization. A single function alone cannot produce the desired results. Management involves continuous planning, organizing, directing and controlling.

(2) **Management is an art as well as science**: Management is an art in the sense of possessing managing skill by a person. Management is science because certain principles, laws are developed which are applicable in place where group activities are coordinated. This will be discussed in detail later in this chapter.

(3) **Management aims at achieving predetermined objectives:** All organizations have objectives that are laid down. Every managerial activity results in achievement of these

predetermined objectives.

(4) **Management is a factor of production:** An enterprise produce goods or services using resources like land, labour, capital, machines etc. These resources themselves cannot realize the organizations goals. The goals are achieved when these are effectively coordinated by the entrepreneur. In case of small enterprises an individual can do such type of job where as in large enterprises the coordination job is done by management. Therefore, management is a factor of production.

(5) **Management is decision-making:** Decision-making is selecting the best among alternative courses. Decision-making is an important function of a manager. Whatever a manager does, he does it by making decisions. The success or failure of an organization depends upon the quality of decision. A manager must make a right decision at right time.

(6) **Universal application:** The principles and concepts of management are applicable to every type of industry. The practice of management is different from one organization to another according to their nature.

(7) **Management is needed at all levels:** The functions of management are common to all levels of organization. The functions of planning, organizing, directing, controlling, decision-making are performed by top level as well as lower level supervisors.

(8) **Management aims at maximum profit:** The resources are properly utilized to maximize profit. Maximizing the profit is the economic function of a manager.

(9) **Dynamic:** Management is not static. Over a period of time new principles, concepts and techniques are developed and adopted by management. Management is changed accordingly to the social change.

Management as a career: Today management is developed as a career focused on specialization. Marketing management, finance management, personal management, industrial management, production management, quality management are some of the specializations in management. Specialists are appointed at various positions of the organizational hierarchy. Hence management is career.

(1) **Management is a profession:** Management is a profession because it possesses the qualities of a profession. The knowledge is imported and transferred. The established principles of management are applied in practice. This is discussed in detail later in this chapter.

(2) **Management is a discipline:** Discipline refers to the field of study having well defined concepts and principles. Classifying management as disciplines implies that it is an accumulated body of knowledge that can be learnt. Thus, management is a subject with principles and concepts. The purpose of studying management is to learn how to apply these principles and concepts at right circumstances, at the right time to produce desired result.

Meaning of Principles of Management: – Principles of management is what a **manager does within the organization**. All the activities whether related to decision making or implementation of such decisions are performed by the managers with the purpose of achieving common goals. Principles of Management is designed to meet the **scope and sequence requirements** of the introductory course on **management**.

<u>What are 14 Principles of Management by Henri Fayol?</u>

1. **Division of Work: –** Dividing the full work of the organization among individuals and creating departments is called the division of work. Division of work leads to specialization, and specialization helps to increases efficiency and efficiency which results in improvements in the productivity and profitability of the organization.
2. **Balancing** <u>Authority and Responsibility</u>**: –** Authority must be equal to Responsibility. According to Henri Fayol, there should be a balance between Authority (Power) and Responsibility (Duties). The right to give orders should not be considered without reference to responsibility. If the authority is more than responsibility then chances are that a manager may misuse it. If responsibility is more than authority then he may feel frustrated.

3. **Discipline: –** Discipline means respect for the rules and regulations of the organization. Discipline may be Self-discipline, or it may be Enforced discipline. No slacking or bending of rules is allowed in any organization. The works must respect the rules that run the organization. To establish discipline, good supervision and impartial judgment are needed.
4. **Unity of Command: –** According to this principle, a subordinate (employee) must have and receive orders from only one superior (boss or manager). To put it another way, a subordinate must report to only one superior. It helps in preventing dual subordination. This decreases the possibilities of "Dual subordination" which creates a problem is a function of managers.
5. **Unity of Direction: –** Unity of direction means activities aimed at the same objective should be organized so that there are one plan and one person in charge. All activities which have the same objective must be directed by one manager, and he must use one plan. This is called the Unity of Direction. For example, all marketing activities such as advertising, sales promotion, pricing policy, etc., must be directed by only one manager.
6. **Subordination of Individual Interests to the General Interest: –** The interest of one individual or one group should not prevail over the general good. The individual interest should be given less importance, while the general interest should be given the most importance. If not, the organization will collapse. The interest of the organizational goal should not be sabotaged by the interest of an individual or on the group.
7. **Remuneration: –** Remuneration is the price for services received. Pay should be fair to both the employee and the firm. If an organization wants efficient employees and best performance, then it should have a good remuneration policy. This policy should give maximum satisfaction to both employers and employees. It should include both financial and non-financial incentives. Remuneration should be based on a systematic attempt to reward good performance.

(3) **Centralization: –** In any company, the management or any authority responsible for the decision-making process should be neutral. However, this depends on the size of

8. an organization. Henri Fayol stressed on the point that there should be a balance between the hierarchy and division of power.
9. **Scalar Chain: –** The chain of command, sometimes called the scalar chain, is the formal line of authority, communication, and responsibility within an organization. The chain of command is usually depicted on an organizational chart, which identifies the superior and subordinate relationships in the organizational structure. Or it is the line of authority from top to bottom of the organization. This chain implements the unity-of-command principle and allows the orderly flow of information. Under the unity of command principle, the instructions flow downward along the chain of command and accountability flows upward.
10. **Order: –** A company should maintain a well-defined work order to have a favourable work culture. The positive atmosphere in the workplace will boost more positive productivity. There must be an orderly placement of the resources such as Men and Women, Money, Materials, etc. Human and material resources must be in the right place at the right time. Misplacement will lead to misuse and disorder.
11. **Equity: –** All employees should be treated equally and respectfully. It's the responsibility of a manager that no employees face discrimination. It creates loyalty and devotion in the employees toward the organization. The equity principle suggests that the managers must be kind as well as equally fair to the subordinates.
12. **Stability of Tenure of Personnel: –** An employee delivers the best if they feel secure in their job. It is the duty of the management to offer job security to their employees. The employees should have job security because instability leads to inefficiency. Successful firms usually had a stable group of employees.
13. **Initiative: –** The management should support and encourage the employees to take initiatives in an organization. It will help them to increase their interest and make then worth. Management should encourage initiative. That is, they should encourage the employees to make their own plans and to execute these plans. This is because an initiative gives satisfaction to the employees and brings success to the organization.
14. **Esprit De Corps: –** Esprit de Corps means "Team Spirit". Therefore, the management should create unity, co-operation, and team-spirit among the employees. They should avoid dividing and rule policy. Harmony, cohesion among personnel. It's a great source of strength in the organization. It is a

quality in every successful business. It is the responsibility of the management to motivate their employees and be supportive of each other regularly. Developing trust and mutual understanding will lead to a positive outcome and work environment.

Foyal presented 14 principles of management as general guides to the management process and management practice. His principles of management are as follows:

15. **Division of work:** This is the principle of specialization which is so well expressed by economists as being necessary to efficiency in the utilization of labour. Fayol goes beyond shop labour to apply the principle to all kinds of work, managerial as well as technical.
16. **Authority and responsibility:** In this principle, Fayol finds authority and responsibility to be related with the letter, the corollary of the former and arising from the latter. The conceives of authority as a combination of official authority

    deriving from a manager's official position and personal authority, "Compounded of intelligence, experience, moral worth, past services etc".
17. **Discipline:** Holding that discipline is "respect for agreements which are directed at achieving obedience, application, energy and the outward marks of respect". Fayol declares that discipline requires good superiors at all levels, clear and fair agreement, and judicious application of penalties.
18. **Unity of command:** This is the principle that an employee should receive orders from one superior only.
19. **Unity of direction:** According to Fayol, unity of direction is the principle thateach group of activities having the same objective must have one head and one plan. As distinguished from the principle of unity of command, Fayol perceives unity of direction as related to the functioning of personnel.
20. **Subordination of individual interest to general interest:** In any group the interest of the group should supersede that of the individual; when these are found to differ, it is the function of management to reconcile them.
21. **Remuneration of personnel:** Fayol perceives that remuneration and methods of payment should be fair and afford the maximum satisfaction to employee and employer.
22. **Centralization:** Although Fayol does not use the term 'Centralization of Authority', his principle definitely refers to the extent to which authority is concentrated or dispersed in an enterprise. Individual circumstances will determine the degree of centralization that will give the best overall yield.
23. **Scalar chair:** Fayol thinks of the scalar chair as a line of authority, a 'Chain of Superiors" from the highest to the lowest ranks and held that, while it is an error of subordinate to depart 'needlessly' from lines of authority, the chain should be short-circuited when scrupulous following of it would be detrimental.
24. **Order:** Breaking this principle into 'Material order' and 'Social Order', Fayol thinks of it as the simple edge of "a place for everything (everyone), and everything (everyone) in its (his) place". This is essentially a principle of organization in the arrangement of things and persons.
25. **Equity:** Fayol perceives this principle as one of eliciting loyalty and devotion from personnel by a combination of kindliness and justice in managers dealing with subordinates.
26. **Stability of tenure of personnel:** Finding that such instability is both the cause and effect of bad management, Fayol points out the dangers and costs of unnecessary turnover.
27. **Initiative:** Initiative is conceived as the thinking out and execution of a plan. Since it is one of the "Keenest satisfactions for an intelligent man to experience", Fayol exhorts managers to "Sacrifice Personal Vanity" in order to permit subordinates to exercise it.

What are Fayol's Five Functions of Management?

1. **Planning: –** Planning means thinking in advance. Planning relates to setting goals, objectives and targets and also describe a mechanism to achieve them at various levels throughout the organization.
2. **Organizing: –** It means the arrangement of resources to do different tasks in the organization. Organizing involves assigning tasks, grouping tasks into departments, delegating authority, and allocating resources across the organization.
3. **Staffing: –** It means the appointment of the right person at the right place or job. Staffing is the process of finding the right worker with appropriate qualifications or experience and recruiting them to fill a job position or role.
4. **Directing: –** It includes the instructions, guidance and motivation given by the manager to their employees and manager adopt their leadership tasks.
5. Controlling**: –** Control is concerned with measuring and minimizing the difference between planned performance and actual performance, and monitoring performance as well as taking corrective actions wherever necessary. Thus management is a complex, integrated and ongoing dynamic process.

What is the importance of management principles?

1. **Optimal Use of Resources: –** The manager eliminates wastage of resources that leads to efficiency in all business functions.
2. **Effective** Leadership **and** Motivation**: –** It helps employees to work in harmony and achieve goals in a coordinated manner. It provides effective leadership and motivation for employees to work hard.
3. **Sound Industrial Relations: –** A manager maintains a balance between employees' demands and organizational needs, helping to reduce industrial disputes.
4. **Achievement of Goals: –** It helps in realizing goals with maximum efficiency by minimizing unnecessary deviations, overlapping efforts and waste motions.
5. **Improvement in Living Standards: –** It improves the standard of living of people by ensuring optimum utilization of scarce resources and survival of the firm in a dynamic environment.
6. **Establishing a Sound Organization: –** Managers help to establish a sound organization through effective authority and responsibility relationships.
7. **Reduces Cost: –** A manager uses cost-effective techniques to reduce production costs and increase production.

1.2 MANAGEMENT FUNCTIONS OR THE PROCESS OF MANAGEMENT

A function is a group of similar activities. There is divergence of view on "What functions are undertaken by managers in organizations?" Some management experts classify these functions into four types and others classify into five types and some others classify them as seven items. The Table 1.1 presents the management functions identified by various writers.

	Writers	**Management Functions**
1	Henry Fayol	Planning, organizing, commending, coordinating, controlling
2	Luther Gulick	Planning, organizing, staffing, directing, coordinating, reportingand budgeting (POSDCORB)
3	Lyndall Urwick	Planning, organizing, commanding, coordinating, communi-cating, forecasting, and investigating.
4	E.F.L. Brech	Planning, organizing, motivating, coordinating, controlling
5	Koontz and O'Donnell	Planning, organizing, staffing, directing (leading), controlling.

(1) **Planning:** Planning is the primary function of management. It is looking ahead and preparing for the future. It determines in advance what should be done. It is conscious

determination of future course of action. This involves determining why to take action? What action? How to take action? When to take action? Planning involves determination of specific objectives, programs, setting policies, strategies, rules and procedures and preparing budgets. Planning is a function which is performed by managers at all levels – top, middle and supervisory. Plans made by top management for the organization as a whole may cover periods as long as five to ten years, whereas plans made by low level managers cover much shorter periods. This "Planning" is discussed in detailed in Chapter–2.

(2) **Organizing:** Organizing is the distribution of work in group-wise or section- wise for effective performance. Once the managers have established objectives and developed plans to achieve them, they must design and develop a human organizationthat will be able to carry out those plans successfully. Such duties in the form of positions, grouping of various positions into departments and sections, assigning duties to individual positions and delegating authority to each position so that the work is carried out as planned.

According to Koonz O'Donnel, "**Organization consists of conscious coordination of people towards a desired goal**". One has to note that different objectives require different kinds of organization to achieve them. For example, an organization forscientific research will have to be very different from one manufacturing automobiles.

(3) **Staffing:** Staffing involves managing various positions of the organizational structure. It involves selecting and placing the right person at the right position. Staffing includes identifying the gap between manpower required and available, identifying the sources from where people will be selected, selecting people, training them, fixing the financial compensation and appraising them periodically. The success of the organization depends upon the successful performance of staffing function.

(4) **Directing:** Planning, organizing and staffing functions are concerned with the preliminary work for the achievement of organizational objectives. The actual performance of the task starts with the function of direction. This function can be called by various names namely "leading", "directing", "motivating", "activating" and so on. Directing involves these sub functions:

(a) **Communicating:** It is the process of passing information from one person toanother.
(b) **Leading:** It is a process by which a manager guides and influences the work of his subordinates.
(c) **Motivating:** It is arousing desire in the minds of workers to give their best to the enterprise.

(5) **Controlling:** Planning, organizing, staffing and directing are required to realize organizational objectives. To ensure that the achieved objectives confirm to the pre- planned objectives control function is necessary. Control is the process of checking to determine whether or not proper progress is being made towards the objectives and goals and acting if necessary to correct any deviations. Control involves three elements:

(a) Establishing standards of performance.
(b) Measuring current performance and comparing it against the established standard.
(c) Taking action to correct any performance that does not meet those standards.

1.3 LEVELS OF MANAGEMENT

People in an organization are arranged in an hierarchy and they all have the relationship of superior-subordinates. Every manager in an organization performs all five management functions. The relative importance of these functions varies along the managerial levels. There may be as many levels in the organization as the number of superiors in a line of command. Some of these levels are merged into one on the basis

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(d) Establishing standards of performance.

(e) Measuring current performance and comparing it against the established standard.

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1.4 LEVELS OF MANAGEMENT

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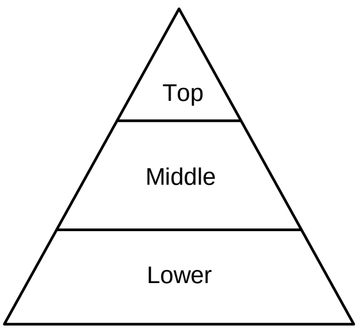

Board of Directors, Chairman, Chief Executive

Department Heads, Divisional Heads, Section Heads

Senior Supervisor, Front Line Supervisors

Fig. 1.3: Levels of management

Top management of an organization consists board of directors, chairman and chief executive officer. Top level management determines goals and objectives. It performs overall planning, organizing, staffing, directing and controlling. It integrates organization with environment, balances the interest groups and is responsible for overall results. Middle management stands between top management and supervisory management level. Middle level management establishes programs for department and carries out functionsfor achieving specific goals. The other functions of middle level management are training and development of employees, integrating various parts of the department. Supervisory management is concerned with efficiency in using resources of the organization. A supervisor is an executor of policies and procedures making a series of decisions with well-defined and specified premises.

UNIT-5

FUNCTIONAL AREAS OF MANAGEMENT

a) Production management
- Functions, Activities
- Productivity
- Quality control
- Production Planning and control

b) Inventory Management
- Need for Inventory management
- Models/Techniques of Inventory management

c) Financial Management
- Functions of Financial management
- Management of Working capital
- Costing (only concept)
- Break even Analysis
- Brief idea about Accounting Terminologies: Book Keeping, Journal entry, Petty Cash book, P&L Accounts, Balance Sheets(only Concepts)

d) Marketing Management
- Concept of Marketing and Marketing Management
- Marketing Techniques (only concepts)
- Concept of 4P s (Price, Place, Product, Promotion)

e) Human Resource Management
- Functions of Personnel Management
- Manpower Planning, Recruitment, Sources of manpower, Selection process, Method of Testing, Methods of Training & Development, Payment of Wages

Meaning of Production Management:

Production Management refers to the application of management principles to the production function in a factory. In other words, production management involves application of planning, organizing, directing and controlling the production process.

Production Management is not independent of marketing, financial and personnel management due to which it is very difficult to formulate some single appropriate definition of Production Management.

Production management, or operations management, focuses on achieving a smooth production process with efficient planning and control of business operations.

The function of production management is to find the perfect balance:

- Right quality
- Right quantity
- Right time
- Right cost

The application of management to the field of production has been the result of at least three developments:

(i) First is the development of factory system of production. Until the emergence of the concept of manufacturing, there was no such thing as management as we know it. It is true that people operated business of one type or another, but for the most part, these people were owners of business and did not regard themselves as managers as well,

(ii) Essentially stems from the first, namely, the development of the large corporation with many owners and the necessity to hire people to operate the business,

(iii) Stems from the work of many of the pioneers of scientific management who were able to demonstrate the value, from a performance and profit point of view, of some of the techniques they were developing.

The following definitions try to explain main characteristics of production management:

(i) In the words of Mr, E.L. Brech:

"Production Management is the process of effective planning and regulating the operations of that section of an enterprise which is responsible for the actual transformation of materials into finished products." This definition limits the scope of production management to those activities of an enterprise which are associated with the transformation process of inputs into outputs. & the definition does not include the human factors involved in a production process. It lays stress on materialistic features only.

(ii) Production Management deals with decision-making related to production process. So that the resulting goods and services are produced in accordance with the quantitative specifications and demand schedule with minimum cost.

According to this definition design and control of the production system are two main functions of production management.

(iii) Production Management is a set of general principles for production economies, facility design, job design, schedule design, quality control, inventory control, work study and cost and budgetary control. This definition explains the main areas of an enterprise where the principles of production management can be applied. This definition clearly points out that production management is not a set of techniques.

It is evident from above definitions that production planning and its control are the main characteristics of production management. In the case of poor planning and control of production activities the organization may not be able to attain its objectives and may result in loss of customer's confidence and retardation in the progress of the establishment.

In short, the main activities of production management can be listed as:

(i) Specification and procurement of input resources namely management, material, and land, labour, equipment and capital.

(ii) Product design and development to determine the production process for transforming the input factors into output of goods and services.

(iii) Supervision and control of transformation process for efficient production of goods and services.

In modern times production management has to perform a variety of functions, namely:

(i) Design and development of production process.

(ii) Production planning and control.

(iii) Implementation of the plan and related activities to produce the desired output.

(iv) Administration and co-ordination of the activities of various components and departments responsible for producing the necessary goods and services.

However, the responsibility of determining the output characteristics and the distribution strategy followed by an organization including pricing and selling policies are normally outside the scope of Production Management.

Functions of production mangement

Production management attempts to utilize 6M’s: Men, Machines, Money, Methods, Materials, and Market in order to better serve consumer needs. Its fundamental goal is to produce products and services in the right quantity, quality, on a schedule, and for optimum money. Production management makes it simple to adopt various technologies and innovative changes in the workplace. Production management is in charge of supervising and controlling all employees involved in the company’s production processes in order to ensure that the target output is achieved.

1. Selection of product and design

Production management helps the organisation select the right product for production and also choose a relevant design for the product. This becomes imperative for the survival of organisations to possess a good understanding of their consumers in order to create products that fully satisfy needs. Products need to go through a detailed evaluation in order to meet customer needs while also remaining cost-efficient.

2. Production planning and control

Choosing the correct production processes for a particular product also becomes important. Decisions must be taken in order to choose the correct type of machines and technology, the capital investment required, and so on. It entails planning prior to production. Decisions like the quantity of production, the flow of processes, and so on are all planned out. Routing is the term used for charting out the sequence of operations for a smooth workflow.

Production control is overseen by the production manager. The actual process is compared and contrasted with the blueprint in place so that all necessary diversions from th original plan can be mapped out and any loopholes in the original plan can be spotted and corrected.

Scheduling is done to set up benchmarks as to when starting and when to complete a particular production activity. Inventory and cost control also need to be taken care of. The allocation of materials, labour, and other processes is called the production schedule.

3. Machine maintenance and replacement

Production management takes care of the maintenance and replacement of machines and equipment to ensure the efficient and smooth working of production processes. This is taken care of by the production manager and the team to prevent speed breaks and halts in production.

Importance of production management

1. Efficent use of capital and resources

Production management minimizes the cost of production and enhances the use of resources to the fullest. A concise blueprint enables proper use of resources and time, minimising disparity between production process and output. Evaluation of production processes and maintenance downtime will ensure processes can be managed efficiently optimising workforce efficiency. A well-thought-out production function will result in high-quality products, a faster rate of production, and a lower cost per unit.

2. Competitive edge

Production management can be a great tool for organisations facing competition in the market. A smoother flow of processes increases efficiency whilst also allowing the company to provide quality products and services. Production management techniques play a role in the effective innovation of new products and facilitate research in developing new and quality products. It can aid organisations in emerging as market leaders since less time spent for production processes means more resources to spare for other domains that may need more attention.

3. Minimizes risk of product failures

Preparing a lucid roadmap and collating information and assumptions helps assess the market and reduce chances of failure. Knowing the requirements and needs of the market will help reduce the chance that a product will flop. Ultimately, product management, like everything else, cannot guarantee success, but it does reduce it.

Scope of Production Management:

The scope of production management is indeed vast. Commencing with the selection of location, production management covers such activities as acquisition of land, constructing building, procuring and installing machinery, purchasing and storing raw materials and converting them into saleable products. Added to the above are other related topics such as quality management, maintenance management, production planning and control, methods improvement and work simplification and other related areas.

UNIT-6

LEADERSHIP AND MOTIVATION

a) Leadership

- Definition and Need/Importance
- Qualities and functions of a leader
- Manager Vs Leader
- Style of Leadership (Autocratic, Democratic, Participative)

b) Motivation

- Definition and characteristics
- Importance of motivation
- Factors affecting motivation
- Theories of motivation (Maslow)
- Methods of Improving Motivation
- Importance of Communication in Business
- Types and Barriers of Communication

Leadership is the potential to influence behaviour of others. It is also defined as the capacity to influence a group towards the realization of a goal. Leaders are required to develop future visions, and to motivate the organizational members to want to achieve the visions.

George R. Terry, "Leadership is a relationship in which one person influences others to work together willingly on related tasks to attain what the leader desires." Koontz and O'Donnell, "Leadership is the process of influencing people so that they will strive willingly towards the achievement of group goals."

According to Keith Davis, "Leadership is the ability to persuade others to seek defined objectives enthusiastically. It is the human factor which binds a group together and motivates it towards goals."

Characteristics of Leadership

1. It is a inter-personal process in which a manager is into influencing and guiding workers towards attainment of goals.
2. It denotes a few qualities to be present in a person which includes intelligence, maturity and personality.
3. It is a group process. It involves two or more people interacting with each other.
4. A leader is involved in shaping and moulding the behaviour of the group towards accomplishment of organizational goals.
5. Leadership is situation bound. There is no best style of leadership. It all depends upon tackling with the situations.

Nature of leadership

- Leadership derives from the power and is similar to, yet distinct from, management. In fact, "leadership" and "management" are different. There can be leaders of completely unorganized groups, but there can be managers only of organized groups. Thus it can be said that a manager is necessarily a leader but a leader may not be a manager.
- Leadership is essential for managing. The ability to lead effectively is one of the keys to being an effective manager because she/he has to combine resources and lead a group to achieve objectives.
- Leadership and motivation are closely interconnected. By understanding motivation, one can appreciate better what people want and why they act as they do. A leader can encourage or dampen workers' motivation by creating a favourable or unfavourable working environment in the organization.
- The essence of leadership is followership. In other words, it is the willingness of people to follow a person that makes that person a leader. Moreover, people tend to follow those whom they see as providing a means of achieving their desires, needs and wants.
- Leadership involves an unequal distribution of power between leaders and group members. Group members are not powerless; they can shape group activities in some ways. Still, the leader will usually have more power than the group members.
- Leaders can influence the followers' behaviour in some ways. Leaders can influence workers either to do ill or well for the company. The leader must be able to empower and motivate the followers to the cause.
- The leader must co-exist with the subordinates or followers and must have a clear idea about their demands and ambitions. This creates loyalty and trust in subordinates for their leader.
- Leadership is to be concerned about values. Followers learn ethics and values from their leaders. Leaders are the real teachers of ethics, and they can reinforce ideas. Leaders need to make positive statements of ethics if they are not hypocritical.
- Leading is a very demanding job both physically and psychologically. The leader must have the strength, power, and ability to meet the bodily requirements; zeal, energy, and patience to meet the mental requirements for leading.

The following points justify the importance of leadership in a concern.

1. **Initiates action-** Leader is a person who starts the work by communicating the policies and plans to the subordinates from where the work actually starts.
2. **Motivation-** A leader proves to be playing an incentive role in the concern's working. He motivates the employees with economic and non-economic rewards and thereby gets the work from the subordinates.
3. **Providing guidance-** A leader has to not only supervise but also play a guiding role for the subordinates. Guidance here means instructing the subordinates the way they have to perform their work effectively and efficiently.

4. **Creating confidence-** Confidence is an important factor which can be achieved through expressing the work efforts to the subordinates, explaining them clearly their role and giving them guidelines to achieve the goals effectively. It is also important to hear the employees with regards to their complaints and problems.
5. **Building morale-** Morale denotes willing co-operation of the employees towards their work and getting them into confidence and winning their trust. A leader can be a morale booster by achieving full co-operation so that they perform with best of their abilities as they work to achieve goals.
6. **Builds work environment-** Management is getting things done from people. An efficient work environment helps in sound and stable growth. Therefore, human relations should be kept into mind by a leader. He should have personal contacts with employees and should listen to their problems and solve them. He should treat employees on humanitarian terms.
7. **Co-ordination-** Co-ordination can be achieved through reconciling personal interests with organizational goals. This synchronization can be achieved through proper and effective co-ordination which should be primary motive of a leader.

Leader versus Manager

> "Leadership and managership are two synonymous terms" is an incorrect statement. Leadership doesn't require any managerial position to act as a leader. On the other hand, a manager can be a true manager only if he has got the traits of leader in him. **By virtue of his/her position, manager has to provide leadership to his group**.

A manager has to perform all five functions to achieve goals, i.e., Planning, Organizing, Staffing, Directing, and Controlling. Leadership is a part of these functions.

Leadership as a general term is not related to managership. A person can be a leader by virtue of qualities in him. For example: leader of a club, class, welfare association, social organization, etc. Therefore, it is true to say that, "All managers are leaders, but all leaders are not managers."

A leader is one who influences the behaviour and work of others in group efforts towards achievement of specified goals in a given situation. On the other hand, manager can be a true manager only if he has got traits of leader in him. Manager at all levels are expected to be the leaders of work groups so that subordinates willingly carry instructions and accept their guidance. A person can be a leader by virtue of all qualities in him.

Leaders and Managers can be compared on the following basis:

Basis	**Manager**	**Leader**
Origin	A person becomes a manager by virtue of his position.	A person becomes a leader on basis of his personal qualities.
Formal Rights	Manager has got formal rights in an organization because of his status.	Rights are not available to a leader.
Followers	The subordinates are the followers of managers.	The group of employees whom the leaders leads are his followers.
Functions	A manager performs all five functions of management.	Leader influences people to work willingly for group objectives.
Necessity	A manager is very essential to a concern.	A leader is required to create cordial relation between person working in and for organization.
Stability	It is more stable.	Leadership is temporary.
Mutual Relationship	All managers are leaders.	All leaders are not managers.
Accountability	Manager is accountable for self and subordinates behaviour and performance.	Leaders have no well defined accountability.
Concern	A manager's concern is organizational goals.	A leader's concern is group goals and member's satisfaction.
Followers	People follow manager by virtue of job description.	People follow them on voluntary basis.
Role continuation	A manager can continue in office till he performs his duties satisfactorily in congruence with	A leader can maintain his position only through day to

	organizational goals.	day wishes of followers.
Sanctions	Manager has command over allocation and distribution of sanctions.	A leader has command over different sanctions and related task records. These sanctions are essentially of informal nature.

Qualities of a Leader

1. **Physical appearance-** A leader must have a pleasing appearance. Physique and health are very important for a good leader.
2. **Vision and foresight-** A leader cannot maintain influence unless he exhibits that he is forward looking. He has to visualize situations and thereby has to frame logical programmes.
3. **Intelligence-** A leader should be intelligent enough to examine problems and difficult situations. He should be analytical who weighs pros and cons and then summarizes the situation. Therefore, a positive bent of mind and mature outlook is very important.
4. **Communicative skills-** A leader must be able to communicate the policies and procedures clearly, precisely and effectively. This can be helpful in persuasion and stimulation.
5. **Objective-** A leader has to be having a fair outlook which is free from bias and which does not reflects his willingness towards a particular individual. He should develop his own opinion and should base his judgement on facts and logic.
6. **Knowledge of work-** A leader should be very precisely knowing the nature of work of his subordinates because it is then he can win the trust and confidence of his subordinates.
7. **Sense of responsibility-** Responsibility and accountability towards an individual's work is very important to bring a sense of influence. A leader must have a sense of responsibility towards organizational goals because only then he can get maximum of capabilities exploited in a real sense. For this, he has to motivate himself and arouse and urge to give best of his abilities. Only then he can motivate the subordinates to the best.
8. **Self-confidence and will-power-** Confidence in himself is important to earn the confidence of the subordinates. He should be trustworthy and should handle the situations with full will power.

Role of a Leader

1. **Required at all levels-** Leadership is a function which is important at all levels of management. In the top level, it is important for getting co-operation in formulation of plans and policies. In the middle and lower level, it is required for interpretation and execution of plans and programmes framed by the top management. Leadership can be exercised through guidance and counselling of the subordinates at the time of execution of plans.
2. **Representative of the organization-** A leader, i.e., a manager is said to be the representative of the enterprise. He has to represent the concern at seminars, conferences, general meetings, etc. His role is to communicate the rationale of the enterprise to outside public. He is also representative of the own department which he leads.
3. **Integrates and reconciles the personal goals with organizational goals-** A leader through leadership traits helps in reconciling/ integrating the personal goals of the employees with the organizational goals. He is trying to co-ordinate the efforts of people towards a common purpose and thereby achieves objectives. This can be done only if he can influence and get willing co-operation and urge to accomplish the objectives.
4. **He solicits support-** A leader is a manager and besides that he is a person who entertains and invites support and co-operation of subordinates. This he can do by his personality, intelligence, maturity and experience which can provide him positive result. In this regard, a leader has to invite

suggestions and if possible implement them into plans and programmes of enterprise. This way, he can solicit full support of employees which results in willingness to work and thereby effectiveness in running of a concern.

5. **As a friend, philosopher and guide-** A leader must possess the three dimensional traits in him. He can be a friend by sharing the feelings, opinions and desires with the employees. He can be a philosopher by utilizing his intelligence and experience and thereby guiding the employees as and when time requires. He can be a guide by supervising and communicating the employees the plans and policies of top management and secure their co-operation to achieve the goals of a concern. At times he can also play the role of a counsellor by counselling and a problem-solving approach. He can listen to the problems of the employees and try to solve them.

Leadership and Motivation

Motivation is a goal-oriented characteristic that helps a person achieve his objectives. It pushes an individual to work hard at achieving his or her goals. An executive must have the right leadership traits to influence motivation. However, there is no specific blueprint for motivation.

As a leader, one should keep an open perspective on human nature. Knowing different needs of subordinates will certainly make the decision-making process easier.

Both an employee as well as manager must possess leadership and motivational traits.

An effective leader must have a thorough knowledge of motivational factors for others. He must understand the basic needs of employees, peers and his superiors. Leadership is used as a means of motivating others.

Given below are important guidelines that outline the basic view of motivation:

- **Harmonize and match the subordinate needs with the organizational needs**. As a leader, the executive must ensure that the business has the same morals and ethics that he seeks in his employees. He should make sure that his subordinates are encouraged and trained in a manner that meets the needs of the business.
- **Appreciation and rewards** are key motivators that influence a person to achieve a desired goal. Rewarding good/ exceptional behavior with a small token of appreciation, certificate or letter can be a great motivator. If a certificate is awarded to a person, it should mention the particular act or the quality for which the individual is being rewarded.
- **Being a role model** is also a key motivator that influences people in reaching their goals. A leader should set a good example to ensure his people to grow and achieve their goals effectively.
- **Encouraging individuals** to get involved in planning and important issues resolution procedure not only motivates them, but also teaches the intricacies of these key decision-making factors. Moreover, it will help everyone to get better understanding of their role in the organization. The communication will be unambiguous and will certainly attract acknowledgement and appreciation from the leader.
- **Developing moral and team spirit** certainly has a key impact on the well-being of an organization. The metal or emotional state of a person constitutes his or her moral fabric. A leader’s actions and decisions affect the morale of his subordinates. Hence, he should always be aware of his decisions and activities. Team spirit is the soul of the organization. The leader should always make sure his subordinates enjoy performing their duties as a team and make themselves a part of the organization’s plans.
- A leader should **step into the shoes of the subordinates** and view things from subordinate’s angle. He should empathize with them during difficult times. Empathizing with their personal problems makes them stronger-mentally and emotionally.

- A meaningful and challenging job accomplished inculcates a sense of achievement among employees. The **executive must make their employees feel they are performing an important work** that is necessary for the organization's well-being and success. This motivational aspect drives them to fulfill goals.

Remember, "**To become an efficient leader, you must be self-motivated**". You must know your identity, your needs and you must have a strong urge to do anything to achieve your goals. Once you are self-motivated, only then you can motivate others to achieve their goals and to harmonize their personal goals with the common goals of the organization.

Types of Leadership Styles

German-American psychologist Kurt Lewin is credited with branding the basic leadership styles in 1939. Lewin and his researchers tasked schoolchildren with an arts and crafts project while the team observed behaviors and responses to different styles of leadership. The idea was to determine which style was most effective to use in business.

Lewin identified three styles of leadership: **Autocratic, Democratic, and Laissez-faire**. Over time, more leadership styles have emerged, and one that is commonly grouped in with Lewin's three is Transformational Leadership. Today, there are a variety of leadership styles in business, but the four primary leadership styles you'll be exposed to include:

- Autocratic
- Democratic
- Laissez-faire
- Transformational

Autocratic Leadership

Also known as Authoritarian Leadership, an Autocratic Leadership style can easily be summed up by the command, "Do as you're told."

In Autocratic Leadership, direction comes from the top, a singular figure who leads a company or team. An autocratic leader determines strategy, policies, procedures, and the direction of the organization, dictating everything to subordinates. Authoritarian leaders are not focused on collaboration with those in their circle, they are rarely interested in feedback, and they prefer to hold all of the power and be in charge.

In Lewin's research, he witnessed four behaviors and results of the Authoritarian Leadership style:

1. All determination of policy is by the leader.
2. Techniques and activity steps are dictated by the authority, one at a time, so that future steps or techniques are uncertain.
3. The leader usually dictates the particular work task and work companions of each member.

4. The dominator is “personal” in his praise and criticism of the work of each team member, but is aloof from active group participation, except when demonstrating. He is friendly or impersonal rather than openly hostile.
   Examples of autocratic leaders include Martha Stewart, and, New England Patriots Head Coach Bill Belichick.

   Qualities: Autocratic leaders often possess qualities subordinates look up to, such as decisiveness, self-confidence, and a steadfast, focused commitment to the goal.
   When It Works: While Autocratic Leadership and the “Don’t question my commands” approach doesn’t sound like fun, there are occasions when this leadership style can be appropriate and effective, such as in urgent or chaotic situations that require someone to reign in the team and make a prompt and effective decision.
   When It Doesn’t Work: Employees working under an autocratic leader may feel micromanaged, and because all directives come from one person at the top, they may become dependent on that person for direction and incapable of making business decisions on their own.
   Democratic Leadership

   Also known as Participative Leadership, the Democratic Leadership approach involves gathering input from your subordinates and team members so everyone has a chance to contribute to the decision-making process. Democratic leaders are still the decision-makers, but their approach allows others to feel engaged and have a stake in the final outcome. Democratic leaders excel at sparking creativity among subordinates, and projects are enhanced when positive contributions come from all sides.

   Lewin’s research into the Democratic Leadership style uncovered these four behaviors and results:

1. All policies were a matter of group discussion and decision, encouraged and assisted by the leader.

2. Activity perspective was gained during the first discussion period. General steps to group goal were sketched, and where technical advice was needed, the leader suggested two or three alternative procedures from which a choice could be made.

3. The members were free to work with anyone, and division of tasks was left to the group leader.

4. The leader was “objective” or “fact-minded” in his praise and criticism and tried to be a regular group member in spirit without doing too much of the work.
   Microsoft Corporation co-founder Bill Gates, Nelson Mandela, and Walt Disney, President of The Walt Disney Company, are often cited as examples of democratic leaders.

   Qualities: Democratic leaders possess a curiosity that drives their desire for input from all sides. Their desire for a participative environment also makes them great communicators, and subordinates often find them easily approachable.
   When It Works: A democratic approach to leadership yields positive results when the group feels part of the solution. This leads to more creative input and often heightens employee morale.
   When It Doesn’t Work: Not every leader achieves success with the democratic approach. Inevitably, there will be employees who feel left out because their ideas or solutions were not chosen. Further, placing confidence in the group to posit solutions can be problematic if the group is not skilled or trained to answer the call.
   Laissez-faire Leadership

   Don’t let the name deceive you. Laissez-faire Leadership is not a “Who cares?” approach. Rather, it involves empowering your employees, being hands-off, and trusting them to accomplish the task at hand without constant questions or micromanagement. Laissez-faire leaders leave decisions to their employees, while staying available to provide feedback when necessary.

According to Lewin, Laissez-faire leaders exhibit four common behaviors and results:

1. Complete freedom for group or individual decision, without any leader participation

2. Various materials supplied by leader, who made it clear that he would supply information when asked, but took no other part in work discussions

3. Complete nonparticipation by leader

4. Very infrequent comments on member activities unless questioned, and no attempt to participate or interfere with the course of events
Former U.S. presidents James Buchanan, Herbert Hoover, and Ronald Reagan, as well as industrialist Andrew Mellon, and businessman Warren Buffet are often cited as examples of Laissez-faire leaders.

Qualities: Laissez-faire leaders are excellent at delegating, and they instill confidence in employees when assigning them tasks without oversight. They are capable of providing constructive criticism when needed, and are often seen as trusting, as they willingly place responsibilities in the hands of employees.
When It Works: The Laissez-faire approach often leads to faster decision-making, as employees don't need to ask a higher up for approval. It is also especially effective in scenarios where the employees or groups are already trained and skilled for the task at hand. These workers are fully competent and don't require supervision, and when they're empowered by a Laissez-faire leader, they may feel more accomplished when they complete their task without a guiding hand or directive.
When It Doesn't Work: Hands-off can be problematic when your team doesn't fully understand the mission. Further, Laissez-faire leaders may find that without direction or oversight, employees don't accomplish as much, or anything at all. Unless you have complete confidence in your employees and their collective ability to complete a task without close supervision, you may reconsider this approach.
Transformational Leadership

Transformational Leadership involves developing a grand vision and rallying your employees around it. Under this style, the team is eager to transform and evolve – personally and professionally – in order to achieve the overall goal. With the organization and employees aligned, teams working under a transformational leader are united for the singular cause, and willing to commit their effort, time, and energy to the organization.

<u>Transformational Leadership Qualities:</u>

1. Understanding what needs to change

2. Ability to stimulate the intellect

3. Knack for encouraging participation

4. Talent for genuine communication

5. Loyalty

6. Sense of the bigger picture

7. Personal integrity

8. An inspiring bearing/presence
   When It Works: Transformational Leadership allows employees to see clearly the vision and goals for the organization, and employees remain loyal and productive while working to accomplish their mission. Transformational Leadership creates workplace harmony as everyone is working together, and the lines of communication are always open between employees and leadership.
   When It Doesn't Work: While a fully engaged and committed group is what every leader wants, Transformational Leadership has been known to contribute to burnout as employees work continuously to achieve goals. This approach also requires constant communication and feedback between leadership and the employees, and if those lines are severed in any way, the latter can feel left out of the big picture.

Motivation

Definition and characteristics

What is Motivation?

Motivation is a psychological process through which a person acts or behaves towards a particular task or activity from start to completion. Motivation drives or pushes a person to behave in a particular way at that point in time. When the motivation is positive, a person is happy, energetic, enthusiastic & self-driven to perform the work and when it is negative motivation, person is demoralized, sad, lethargic & pessimistic leading to drop in productivity and

<u>Definition</u>

Motivation is regarded as "the inner state that energizes activities and directs or channels behavior towards the goal".

Motivation is the process that arouses action, sustains the activity in progress and that regulates the pattern of activity.

Entrepreneurial motivation

The entrepreneurial motivation is the process that activates and motivates the entrepreneur to exert higher level of efforts for the achievement of his/her entrepreneurial goals. In other words, the entrepreneurial motivation refers to the forces or drive within an entrepreneur that affect the direction, intensity, and persistence of his / her voluntary behaviour as entrepreneur. So to say, a motivational entrepreneur will be willing to exert a particular level of effort (intensity), for a certain period of time (persistence) toward a particular goal (direction).

<u>Nature of Motivation</u>

The nature of motivation emerging out of above definitions can be expressed as follows:

1. Motivation is internal to man

Motivation cannot be seen because it is internal to man. It is externalized via behavior. It activates the man to move toward his / her goal.

2. A Single motive can cause different behaviors

A person with a single desire or motive to earn prestige in the society may move towards to join politics, attain additional education and training, join identical groups, and change his outward appearance.

3.Different motives may result in single behaviour

It is also possible that the same or single behaviour may be caused by many motives. For example, if a person buys a car, his such behaviour may be caused by different motives such as to look attractive, be respectable, gain acceptance from similar group of persons, differentiate the status, and so on.

4.Motives come and go

Like tides, motives can emerge and then disappear. Motives emerged at a point of time may not remain with the same intensity at other point of time. For instance, an entrepreneur overly concerned about maximization of profit earning during his initial age as entrepreneur may turn his concern towards other higher things like contributing towards philanthropic activities in social health and education once he starts earning sufficient profits.

5.Motives interact with the environment

The environment in which we live at a point of time may either trigger or suppress our motives. You probably have experienced environment or situation when the intensity of your hunger picked up just you smelled the odour of palatable food.

You may desire an excellent performance bagging the first position in your examination but at the same time may also be quite sensitive to being shunned and disliked by your class mates if you really perform too well and get too much of praise and appreciation from your teachers. Thus, what all this indicates is that human behaviour is the result of several forces differing in both direction and intent.
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Importance of Motivation

It is an important psychological factor for any individual as it defines the work, ambition and drive of that person to do any work. A person with high levels of motivation is motivated to do good quality work, help others, spread their energy and focus on achieving goals. On the contrary, a person with low levels of motivation, demotivates others, works shabbily and creates a negative atmosphere.

It is an important human factor in an individual's personal as well as professional life. Positive motivation for people is required in every field like business, sports, politics, entrepreneurship etc. It is the desire of an individual to work towards a motive, which is a certain task. In business, good motivation helps employees learn important managerial skills like leadership, team management, time management, decision making, communication etc.

4 Types of Motivation

It is driven by several factors which influences the behavior and attitude of an individual. Based on the different factors and the kind of impact it has on a person, there are different types of motivation. The different types of motivation in people are:

1. Intrinsic Motivation

This type comes from within a person to do a task or achieve a particular goal. It is a feeling of being self-driven and achieving objectives for oneself. Intrinsic motivation is driven by motives like social acceptance, eating food, desires to achieve goals, biological needs etc.

2. Extrinsic Motivation

This type drives an individual due to external forces or parameters. Some other person or organization motivates the individual to work hard to achieve certain goals or tasks. Extrinsic motivation is driven by motives like financial bonus, rewards, appreciation, promotion, punishment, demotion etc.

3. Positive Motivation

This type drives an individual by offering positive accolades and rewards for performing a task. In this type of motivation, the individual is rewarded by monetary benefits, promotions etc. which drives an individual to work hard.

4. Negative Motivation

This is where fear and threat are used as a parameter to get the work done. In this type of motivation, individuals are threatened with things like demotion, reducing benefits, withdrawing merits etc.

For every individual, all the types are interlinked based on which he or she takes an action. The resultant behavior of any person is basis the motivation types he or she has been influenced by.

Top Factors Influencing Motivation

Keeping employees motivated is the biggest challenge for companies for ensuring that they give a high productive output at work and help in achieving company goals. A positive motivation amongst employees helps drive the business positively & enhances creativity. On the other hand, a demotivated employee will not contribute efficiently and slowdown progress at workplace. The key elements & top factors which influence employees in business or people in general are as follows:

1. Salary

Monetary compensation & benefits like gross salary, perks, performance bonuses etc. are the biggest motivation factors. The better the salary and monetary benefits, the higher is the motivation level & passion of a person towards a job.

2. Recognition

Rewards, recognition etc. are important for ensuring high enthusiasm levels for an employee. If the hard work of an individual is appreciated, it keeps them motivated to perform better.

3. Work Ethics

Ethical working environment, honesty etc. are important factors for any individual. Good work ethics in a company helps keep employees motivated at work place.

On the other hand, if the work environment is not ethical, then the workforce might be demotivated.

4. Transparency with Leadership

The leadership in an organization helps in employee motivation if there are transparent discussion and flatter hierarchies. The senior management has to ensure that all subordinates are happy, focused & motivated.

5. Culture at Work

A good, vibrant, positive culture at workplace is always an important factor. People from different backgrounds, religions, countries etc. working together helps create a social bond at workplace.

6. Learning and Development

Another factor influencing is the training and development opportunities that a person gets. L&D helps individuals develop more skills and have better opportunities in their professional career.

7. Work Life Balance

Having a good quality of work life (QWL) helps in the motivation of people. A good work life balance ensures that a person can give quality time to both office work as well as family.

8. Career Growth Opportunities

Career development opportunities have a positive influence on the motivation of any person. If a person knows their future & career path is secure, they tend to work with more passion.

9. Health Benefits

Health benefits, insurance & other incentives act as a source of motivation for people. If the medical bills, hospitalization charges etc. are taken care of by the company, it helps build a strong trust.

10. Communication

A positive & transparent communication between managers and subordinates gives a sense of belonging and adds to the employee's motivation. Discussion related to work as well as personal life help make a friendly bond at workplace.

There are not finite factors influencing positive attitude of an employee. These keep on changing depending upon the type of environment, job responsibility, experience in life etc. Hygiene Factors of motivation are also given by the Hygiene Theory.

Advantages of Motivation

There are several advantages of an employee being motivated at work. Some of the main benefits of motivation for employees and companies are as follows:

1. Feeling of belongingness and self-respect.

2. A motivated employee is excited to learn and contribute more.

3. Higher productive output due to positive motivation.

4. Reduced absenteeism and lower attrition.

5. Employees high on motivation help reduce costs & improve profits.

6. Reduce stress & anxiety at workplace.

Disadvantages of Motivation

Despite several advantages, there are some probable drawbacks as well. Some of them are:

1. It is a relative feeling and hence there is no way of ensuring that everyone would feel the same way.

2. Motivating employees in a company requires additional efforts, money and time on employee related activities.

3. It among people is mostly a short-term feeling. Afterwards a person become demotivated or even uninterested about a particular task.

Examples of Motivation

Motivation is relevant in every field of life for any individual. There can be many examples based on type, situation, field of work etc. Some examples are as follows.

1. Consider a person who has joined a university. After a lot of hard work, he or she gets admission in their preferred course. Hence the intrinsic motivation for the individual would be perform well in academics, get a good score, learn new skills and get a job opportunity with a good company. Since this is an internal feeling, the driving motives are achieving a personal goal.

2. An example of extrinsic motivation can be seen from a footballer, cricketer or any other sportsperson. Any sportsperson wants to excel in their sport not only for themselves but to also become a crowd favorite and get their attention. Hence, he or she gets highly motivated by this external factor which drives them to perform better in their sport.

Maslow's Hierarchy of Needs Theory

One of the most popular needs theories is **Abraham Maslow's hierarchy of needs theory**. Maslow proposed that motivation is the result of a person's attempt at fulfilling five basic needs: physiological, safety, social, esteem and self-actualization. According to Maslow, these needs can create internal pressures that can influence a person's behavior.

Physiological needs are those needs required for human survival such as air, food, water, shelter, clothing and sleep. As a manager, you can account for the physiological needs of your employees by providing comfortable working conditions, reasonable work hours and the necessary breaks to use the bathroom and eat and/or drink.

Safety needs include those needs that provide a person with a sense of security and well-being. Personal security, financial security, good health and protection from accidents, harm and their adverse effects are all included in safety needs. As a manager, you can account for the safety needs of your employees by providing safe working conditions, secure compensation (such as a salary) and job security, which is especially important in a bad economy.

Social needs, also called **love and belonging**, refer to the need to feel a sense of belonging and acceptance. Social needs are important to humans so that they do not feel alone, isolated and depressed. Friendships, family and intimacy all work to fulfill social needs. As a manager, you can account for the social needs of your employees by making sure each of your employees know one another, encouraging cooperative teamwork, being an accessible and kind supervisor and promoting a good work-life balance.

Esteem needs refer to the need for self-esteem and respect, with self-respect being slightly more important than gaining respect and admiration from others. As a manager, you can account for the esteem needs of your employees by offering praise and recognition when the employee does well, and offering promotions and additional responsibility to reflect your belief that they are a valued employee.

Self-actualization needs describe a person's need to reach his or her full potential. The need to become what one is capable of is something that is highly personal. While I might have the need to be a good parent, you might have the need to hold an executive-level position within your organization. Because this need is individualized, as a manager, you can account for this need by providing challenging work, inviting employees to participate in decision-making and giving them flexibility and autonomy in their jobs.

Methods of Improving Motivation

1. Lead with vision

Everyone wants to know that their efforts are driving towards something. What's the next step? What does success look like for the company? A destination helps to motivate the journey, so make sure the vision for the company is clear.

2. Make sure everyone understands the 'why'

Your employees will know what needs to be done, but you need to explain further; you need to communicate the 'why' of each task. The why is the company's overall mission. If everyone knows how their individual actions can personally add to the overall goal of the company, it brings much-needed intrinsic motivation to even the most simple task.

3. Set frequent clear targets

You obviously have big targets that you want to hit as a company, but smaller goals are the key to motivation. All goals should add to the overall target, but breaking this into more attainable chunks feels less overwhelming. If employees are frequently hitting targets, the feeling of satisfaction grows and will act as a great motivator to continue on to the next set of goals.

4. Recognise and reward great work

Employees need to know that their managers appreciate their hard work. Giving well-deserved recognition not only increases self-esteem but also enthusiasm and team morale.

A recognition platform is an effective tool to celebrate star performers who embody your company values. It's important to support employees all year round with team-building exercises, wellbeing schemes, secondment opportunities and EAP programs. However, for more special occasions, such as a 10-year work anniversary or employee of the month, we recommend going a step further and rewarding your teams.

The reward doesn't necessarily have to be monetary, it could be a gesture, such as giving them a day off. Rewards are particularly good promoters of motivation and job satisfaction. So if you want to encourage your employees to go the extra mile, make their effort worth it with a gift or thoughtful act.

5. Give your team autonomy

Time is precious. So when we don't feel in control of our time and energy, motivation levels can really drop. Allowing for some elements of freedom in the workplace, whether that's flexible working hours or unlimited time off, demonstrates trust from leaders to employees. This adds motivation, as the satisfaction of a job well done comes with the feeling that they were in control and did it on their terms.

6. Create a welcoming workplace environment

No one wants to sit in a gloomy office and desperately wait for home time every day. If workplaces create a friendly culture, with areas for rest and play, employees will look forward to coming into work. The saying 'work hard, play hard' is important here. As motivation and mood go hand in hand, a poor mood can affect the ability to concentrate and will lower the feeling of energy in the workplace.

7. Offer impressive benefits

Make everyone feel that they are working in the best place they can be. Offering employee benefits and perks, such as the wide range available through our perks platform, and fringe benefits that aim to make your people's lives better both in and outside of work, helps to boost the mood and sense of loyalty to the company. Make sure your benefits suit your employees' unique needs; for instance, if they live in London, they may require a London weighted allowance.

8. Encourage teamwork

Collaboration between teams in the company allows ideas to be developed further. Working with those with different skill sets will, in turn, create more innovative results. In teams, there's power in numbers and anyone experiencing a lack of motivation should be boosted by those around them.

9. Create a career path

No one wants to be static for long. We all want to know that we are going somewhere and focus on that next step. Ask employees what they want from their career, and lay out what they need to do to get there. Have growth conversations with team members to design a career path; this will help to create the drive to reach the next stage and feel that they have a long and productive journey ahead within the company.

10. Support staff motivation by supporting employee wellbeing

When organisations are implementing motivation strategies or employee incentive schemes, they often overlook wellbeing.

There's no dispute that using rewards as an incentive is a great motivator. However, if an employee is feeling tired or overwhelmed they won't produce their best work – no matter how hard they try.

Taking a well-rounded approach to wellbeing, so addressing your teams' mental, emotional, and physical health is a great way of keeping them healthy, happy, and at the top of their game.

Why is communication important in business?

Communication in business is important to convey clear, strong messages about strategy, customer service and branding. A business building a brand reflects a consistent message tailored to its audience. Internal communication builds rapport among employees and managers and encourages teamwork and collaboration.

Good communicators are vital to a business's profit margin and reputation. Employees who communicate well and understand the company's vision help make the company successful. Using effective communication skills can benefit a business and its employees in a variety of ways, including:

1. Building better teams

Effective communication builds a positive atmosphere where teams can flourish. When communication is positive and encouraging, team members become stronger and work better together. Those who practice good communication skills make working beside them easier and less stressful. Job tensions or friction are promptly addressed since effective communicators work toward finding solutions.

2. Preventing misunderstandings and conflicts

Listening is a part of communicating used to understand a situation fully. Good communicators listen well and react appropriately, and they're more likely to find solutions without becoming defensive. When conflicts arise, good communicators address the problem right away and listen to the other side.

3. Improving customer service

Businesses need to know what their customers want and need from them to serve them. Forming business relationships is how companies build trust in their brand and bring customers back. Reliable and positive communication with customers is paramount to maintaining a valued customer service reputation.

Effective communication with customers by answering questions and providing solutions helps to improve the business's reputation. Communicating well over the phone or through social media or email improves customer confidence and loyalty. Written communication that is consistent with the brand's voice can help promote community between customers and organizations.

4. Meeting goals and earning success

For a business to be successful, short- and long-term goals must be clear and concise. When employees know what the goals and visions of their company are, they are better able to focus their efforts on achieving them.

Communicate goals clearly, and regularly remind employees of the team or individual goals to keep them motivated and encouraged. It can also help employees understand the importance of their roles in making the company successful, which can improve loyalty within the company and reduce hiring turnover.

5. Promoting creativity and innovation

People who are good communicators are usually open to sharing ideas without judgment. When trust is built based on effective communication, team members are more apt to share with each other to inspire creativity.

Effective communicators are more inclined to collaborate to find creative solutions, as well. When employees are keen to share ideas, companies can transform their approaches and find unique ways to serve customers and clients.

6. Advancing individual career prospects

Good communication skills, such as open and honest communication, eliminate surprises and reinforce your commitment to solve problems and support the business. This can help you earn promotions and secure leadership positions in your current organization or pursue new advanced career opportunities elsewhere.

Barriers To Effective Communication

The process of communication has multiple barriers. The intended communique will often be disturbed and distorted leading to a condition of misunderstanding and failure of communication. The Barriers to effective communication could be of many types like linguistic, psychological, emotional, physical, and cultural etc. We will see all of these types in detail below.

Linguistic Barriers

The language barrier is one of the main barriers that limit effective communication. Language is the most commonly employed tool of communication. The fact that each major region has its own language is one of the Barriers to effective communication. Sometimes even a thick dialect may render the communication ineffective.

As per some estimates, the dialects of every two regions changes within a few kilometers. Even in the same workplace, different employees will have different linguistic skills. As a result, the communication channels that span across the organization would be affected by this.

Thus keeping this barrier in mind, different considerations have to be made for different employees. Some of them are very proficient in a certain language and others will be ok with these languages.

Psychological Barriers

There are various mental and psychological issues that may be barriers to effective communication. Some people have stage fear, speech disorders, phobia, depression etc. All of these conditions are very difficult to manage sometimes and will most certainly limit the ease of communication.

Emotional Barriers

The emotional IQ of a person determines the ease and comfort with which they can communicate. A person who is emotionally mature will be able to communicate effectively. On the other hand, people who let their emotions take over will face certain difficulties.

A perfect mixture of emotions and facts is necessary for effective communication. Emotions like anger, frustration, humour, can blur the decision-making capacities of a person and thus limit the effectiveness of their communication.

Physical Barriers to Communication

They are the most obvious barriers to effective communication. These barriers are mostly easily removable in principle at least. They include barriers like noise, closed doors, faulty equipment used for communication, closed cabins, etc. Sometimes, in a large office, the physical separation between various employees combined with faulty equipment may result in severe barriers to effective communication.

Cultural Barriers of Communication

As the world is getting more and more globalized, any large office may have people from several parts of the world. Different cultures have a different meaning for several basic values of society. Dressing, Religions or lack of them, food, drinks, pets, and the general behaviour will change drastically from one culture to another.

Hence it is a must that we must take these different cultures into account while communication. This is what we call being culturally appropriate. In many multinational companies, special courses are offered at the orientation stages that let people know about other cultures and how to be courteous and tolerant of others.

Organisational Structure Barriers

As we saw there are many methods of communication at an organizational level. Each of these methods has its own problems and constraints that may become barriers to effective communication. Most of these barriers arise because of misinformation or lack of appropriate transparency available to the employees.

Attitude Barriers

Certain people like to be left alone. They are the introverts or just people who are not very social. Others like to be social or sometimes extra clingy! Both these cases could become a barrier to communication. Some people have attitude issues, like huge ego and inconsiderate behaviours.

These employees can cause severe strains in the communication channels that they are present in. Certain personality traits like shyness, anger, social anxiety may be removable through courses and proper training. However, problems like egocentric behaviour and selfishness may not be correctable.

Perception Barriers

Different people perceive the same things differently. This is a fact which we must consider during the communication process. Knowledge of the perception levels of the audience is crucial to effective communication. All the messages or communique must be easy and clear. There shouldn't be any room for a diversified interpretational set.

Physiological Barriers

Certain disorders or diseases or other limitations could also prevent effective communication between the various channels of an organization. The shrillness of voice, dyslexia, etc are some examples of physiological barriers to effective communication. However, these are not crucial because they can easily be compensated and removed.

Technological Barriers & Socio-religious Barriers

Other barriers include the technological barriers. The technology is developing fast and as a result, it becomes difficult to keep up with the newest developments. Hence sometimes the technological advance may become a barrier. In addition to this, the cost of technology is sometimes very high.

Most of the organizations will not be able to afford a decent tech for the purpose of communication. Hence, this becomes a very crucial barrier. Other barriers are socio-religious barriers. In a patriarchal society, a woman or a transgender may face many difficulties and barriers while communicating.

UNIT-7
Work Culture, TQM & Safety

- Human relationship and Performance in Organization
- Relations with Peers, Superiors and Subordinates
- TQM concepts: Quality Policy, Quality Management, Quality system
- Accidents and Safety, Cause, preventive measures, General Safety Rules ,
- Personal Protection Equipment(PPE)

Human relationship and Performance in Organization

Human relations is critical for developing and maintaining a positive work environment, retaining employees and encouraging productivity. By making human relations the focus of your management approach, you can effectively create a workplace culture in which your employees can thrive. Developing your human relations skills can help you modify your management style so it's human relations-centric.

What is human relations?

Human relations refers to the ability to interact in a healthy manner with others and build strong relationships. From the perspective of managers in a company, it involves the process of creating systems and communication channels to enable group employee relationships as well as strong one-on-one relationships. Additionally, it includes the process of training people for specific roles, addressing their needs, resolving conflicts between employees or between management and employees and creating a positive workplace culture.

Why is human relations important?

Human relations is critical in business for a variety of reasons, including:

- **Employee productivity:** When relationships between a company's managers and its employees are supportive and positive, productivity is shown to increase. Additionally, when employees are dependent upon each other with their work, it directly influences productivity. When employees feel that they're treated with respect and recognized for contributing to a company's success, they are more likely to produce quality work.
- **Motivation:** When employees recognize that performing job-related tasks will allow them to feel appreciated and valued, they are more motivated to do them.
- **Positive perception:** An employee's interpersonal and communication skills impact the perception that other employees have about their ability to positively contribute to the workplace. A positive perception of an employee's work and attitude could create opportunities for future advancement.
- **Employee retention:** Employees who feel they're treated respectfully by their employer are more likely to continue working for them for a longer period of time. By using human relations to establish a feeling of mutual trust and value, a company creates an environment where employees feel that they matter and are less likely to pursue other opportunities.
- **Creativity:** In the workplace, creativity is often dependent on the employee's ability to communicate and share ideas with their colleagues. By creating a positive workplace culture and nurturing workplace relationships and teamwork, a business can encourage creativity and more quickly develop products and services to meets the consumer's changing needs.

Employee Collaboration and Workplace Culture

Human relations in the workplace are a major part of what makes a business work. Emplo yees must frequently work together on projects, communicate ideas and provide motivation to get things done. Without a stable and inviting workplace culture, difficult challenges can arise both in the logistics of managing employees and in the bottom line. Businesses with engaging workplaces and a well-trained workforce are more likely to retain and attract qualified employees, foster loyalty with customers and more quickly adapt to meet the needs of a changing marketplace.

Improving Employee Retention

The quality of workplace relations is critical to employee retention. Employee retention may seem trivial – especially in a workplace that is used to a high turnover – but managers must remember that turnover is financially very costly. Every new employee requires a substantial investment of time and energy in their recruitment and training.

In addition, severing ties with old employees can sometimes be challenging, especially if the circumstances are not particularly amicable. Making sure quality employees remain interested and engaged in the business requires patience, compassion and flexibility, but can actually make the business more financially sound.

Motivation and Productivity

Workplace relationships provide a source of employee motivation, which is important to maintaining productivity. Employees who are interested in their work and in the well-being of other employees tend to be more productive than those who are not. This productivity pays obvious financial dividends to the company, as it can get more done in less time with fewer costs. Building relationships, by both recognizing an employee's value to the company and a concern for their needs, often goes a long way.

Fostering Employee Creativity

The modern business environment often rewards businesses that are able to quickly develop products that meet changing consumer needs. In some industries – such as technology, for example – employees' ability to come up with effective new ideas is often the difference between the entire company's success and failure. Employees' creativity is often dependent on their ability to communicate with other employees and share ideas. Without quality workplace relationships, employees are less likely to be able to develop and share the solutions that a business needs to survive.

TQM concepts: Quality Policy, Quality Management, Quality system

Total Quality Management (TQM) is a management framework based on the belief that an organization can build long-term success by having all its members, from low-level workers to its highest ranking executives, focus on improving quality and, thus, delivering customer satisfaction.

A core definition of total quality management (TQM) describes a management approach to long-term success through customer satisfaction. In a TQM effort, all members of an organization participate in improving processes, products, services, and the culture in which they work.

PRIMARY ELEMENTS OF TQM

TQM can be summarized as a management system for a customer-focused organization that involves all employees in continual improvement. It uses strategy, data, and effective communications to integrate the quality discipline into the culture and activities of the organization. Many of these concepts are present in modern quality management systems, the successor to TQM. Here are the 8 principles of total quality management:

1. **Customer-focused:** The customer ultimately determines the level of quality. No matter what an organization does to foster quality improvement—training employees, integrating quality into the design process, or upgrading computers or software—the customer determines whether the efforts were worthwhile.
2. **Total employee involvement:** All employees participate in working toward common goals. Total employee commitment can only be obtained after fear has been driven from the workplace, when empowerment has occurred, and when management has provided the proper environment. High-performance work systems integrate continuous improvement efforts with normal business operations. Self-managed work teams are one form of empowerment.
3. **Process-centered:** A fundamental part of TQM is a focus on process thinking. A process is a series of steps that take inputs from suppliers (internal or external) and transforms them into outputs that are delivered to customers (internal or external). The steps required to carry out the process are defined, and performance measures are continuously monitored in order to detect unexpected variation.
4. **Integrated system:** Although an organization may consist of many different functional specialties often organized into vertically structured departments, it is the horizontal processes interconnecting these functions that are the focus of TQM.
5.
    1. Micro-processes add up to larger processes, and all processes aggregate into the business processes required for defining and implementing strategy. Everyone must understand the vision, mission, and guiding principles as well as the quality policies, objectives, and critical processes of the organization. Business performance must be monitored and communicated continuously.
    2. An integrated business system may be modeled after the Baldrige Award criteria and/or incorporate the ISO 9000 standards. Every organization has a unique work culture, and it is virtually impossible to achieve excellence in its products and services unless a good quality culture has been fostered. Thus, an integrated system connects business improvement elements in an attempt to continually improve and exceed the expectations of customers, employees, and other stakeholders.
6. **Strategic and systematic approach:** A critical part of the management of quality is the strategic and systematic approach to achieving an organization's vision, mission, and goals. This process, called strategic planning or strategic management, includes the formulation of a strategic plan that integrates quality as a core component.
7. **Continual improvement:** A large aspect of TQM is continual process improvement. Continual improvement drives an organization to be both analytical and creative in finding ways to become more competitive and more effective at meeting stakeholder expectations.
8. **Fact-based decision making:** In order to know how well an organization is performing, data on performance measures are necessary. TQM requires that an organization continually collect and analyze data in order to improve decision making accuracy, achieve consensus, and allow prediction based on past history.
9. **Communications:** During times of organizational change, as well as part of day-to-day operation, effective communications plays a large part in maintaining morale and in motivating employees at all levels. Communications involve strategies, method, and timeliness.

TQM requires organizations to focus on continuous improvement, or kaizen. It focuses on process improvements over the long term, rather than simply emphasizing short-term financial gains.

Principles of TQM

TQM prescribes a series of ways for organizations to accomplish this, with the pathway to successful continuous improvement centered on the use of strategy, data and effective communication to instill a discipline of quality into the organization's culture and processes.

More specifically, TQM puts a spotlight on the processes that organizations use to produce their products, and it calls for organizations to define those processes, continuously monitor and measure their performance, and use that performance data to drive improvements. Furthermore, it calls for all employees, as well as all organizational departments, to be part of this process.

TQM's objectives are to eliminate waste and increase efficiencies by ensuring that the production process of the organization's product (or service) is done right the first time.

This management framework was initially applied to companies in the manufacturing sector, but, over the decades, organizations in other sectors have adopted it, as well.

Quality Policy

The quality policy is where the company has stated its promise. The promise to strive for consistent quality and constant improvement. The quality policy is a short document created by the executive management team.

It displays exactly what quality means to their company specifically. The purpose of their product or service and the strategic direction they are planning to take to achieve these goals.

You must show this document to all employees so all are aware and the company is aligned with their objective. The document is also made public to be accessed by customers and investors for reference and reassurance.

In short, a quality policy describes your business, what it offers and your commitment to delivering the highest quality product. Customer requirements are accounted for in the quality policy. Quality objectives are the goals linked to meeting customer requirements. These objectives are within the quality policy statements.

Three Key Elements

The quality policy should really have three key elements.

1. It should be developed by members of senior management.
2. It should show commitment to quality and customer satisfaction.
3. It should be relatively easy to generate key quality objectives from it.

A Quality Policy is a requirement of any formalized quality management system (QMS), and its purpose is to set the framework for the commitment of the scope of the QMS for internal and external stakeholders.

A Quality Policy is typically a brief statement that aligns with an organization's purpose, mission, and strategic direction. It provides a framework for quality objectives and includes a commitment to meet applicable requirements (ISO 9001, customer, statutory, or regulatory) as well as to continually improve.

But a Quality Policy can be so much more. It can become a way to drive passion for cultural change within an organization by incorporating pieces of it into routine meetings and embedding it into employee objectives and performance measurement.

For example, some companies will incorporate a quality objective directly into an employee's annual goals. Examples could include being involved in a continual improvement project or simply maintaining compliance.

Most companies will post their Quality Policy on their websites and within company buildings as a visual reminder to employees, customers, and suppliers about their commitment to quality.

Here is our Quality Policy that we believe reflects the everyday way of life and commitment to quality practiced at Thermo Fisher Scientific:

Quality Policy

We fulfill our Mission to enable our customers to make the world healthier, cleaner, and safer by continuously improving the quality of our products and services and by ensuring global regulatory compliance.

- Quality for our customers means they can rely on our products and services to consistently meet their specifications and requirements.
- Quality for our colleagues means we take personal ownership to ensure our work meets customer requirements and is error free from design through use.
- Quality for regulatory authorities means that we operate at the highest ethical standards and meet or exceed all applicable regulatory requirements.
- Quality for our company means we drive a continuous improvement culture that is enabled by practical process improvement (PPI) and our company's quality system.

What Is Quality Management?

Quality management is the act of overseeing all activities and tasks that must be accomplished to maintain a desired level of excellence. This includes the determination of a quality policy, creating and implementing quality planning and assurance, and quality control and quality improvement. It is also referred to as total quality management (TQM).

A Quality Management System, or QMS, is essential for improving your business's operations and enabling it to meet the requirements of customers. Many types of quality management systems are built around ISO 9000, a framework requiring all business processes to be documented and for people to follow these closely. Whether you manage all your processes in-house or outsource to, for example, a , a QMS allows your organization to remain compliant with the latest standards and regulations.

Once your company is ISO certified, implementing Total Quality Management (TQM) enables you to evaluate all processes to improve efficiency and reduce waste. TQM can be implemented in all facets of your business, from the front office to the production floor.

In general, quality management focuses on long-term goals through the implementation of short-term initiatives.

- Quality management is the act of overseeing all activities and tasks needed to maintain a desired level of excellence.

- Quality management includes the determination of a quality policy, creating and implementing quality planning and assurance, and quality control and quality improvement.
- TQM requires that all stakeholders in a business work together to improve processes, products, services and the culture of the company itself.

Understanding Quality Management

At its core, TQM is a business philosophy that champions the idea that the long-term success of a company comes from customer satisfaction and loyalty. TQM requires that all stakeholders in a business work together to improve processes, products, services and the culture of the company itself.

While TQM seems like an intuitive process, it came about as a revolutionary idea. The 1920s saw the rise in reliance on statistics and statistical theory in business, and the first-ever known control chart was made in 1924. People began to build on theories of statistics and ended up collectively creating the method of statistical process control (SPC). However, it wasn't successfully implemented in a business setting until the 1950s.1

It was during this time that Japan was faced with a harsh industrial economic environment. Its citizens were thought to be largely illiterate, and its products were known to be of low quality. Key businesses in Japan saw these deficiencies and looked to make a change. Relying on pioneers in statistical thinking, companies such as Toyota integrated the idea of quality management and quality control into their production processes.

By the end of the 1960s, Japan completely flipped its narrative and became known as one of the most efficient export countries, with some of the most admired products. Effective quality management resulted in better products that could be produced at a cheaper price.

Real-World Example of Quality Management

The most famous example of TQM is Toyota's implementation of the Kanban system. A kanban is a physical signal that creates a chain reaction, resulting in a specific action. Toyota used this idea to implement its just-in-time (JIT) inventory process. To make its assembly line more efficient, the company decided to keep just enough inventory on hand to fill customer orders as they were generated.

Therefore, all parts of Toyota's assembly line are assigned a physical card that has an associated inventory number. Right before a part is installed in a car, the card is removed and moved up the supply chain, effectively requesting another of the same part. This allows the company to keep its inventory lean and not overstock unnecessary assets.2

A quality system is a structure for managing the quality of the output of a manufacturer.

We have a very stringent quality system for inspecting items and delivering the best products.
A good quality system prevents errors from occurring rather than correcting them after they have happened.
A quality system is a structure for managing the quality of the output of a manufacturer.

Accidents in the workplace
Your employer has a duty to protect you and tell you about health and safety issues that affect you. They must also report certain accidents and incidents, pay you sick pay and give you time off because of an accident at work should you need it.

Reporting an accident at work

Your employer must report serious work-related accidents, diseases and dangerous incidents to the Health and Safety Executive for Northern Ireland (HSENI) or the Health and Safety department at your local authority. They must report:

- death
- major injuries, a broken arm or ribs for example
- dangerous incidents like the collapse of scaffolding, people overcome by gas
- any other injury that stops an employee from doing their normal work for more than three days
- disease

The reporting must be done by your employer, but if you're involved it's a good idea to make sure it has been reported.

Health and safety at work

Your employer has to carry out a risk assessment and do what's needed to take care of the health and safety of employees and visitors. This includes deciding how many first aiders are needed and what kind of first aid equipment and facilities should be provided.

First aiders have no statutory right to extra pay, but some employers do offer this. Employees must also take reasonable care over their own health and safety.

- Employees' health and safety responsibilities
- Employers' health and safety responsibilities

Recording accidents

Any injury at work - including minor injuries - should be recorded in your employer's 'accident book'.

All employers (except for very small companies) must keep an accident book. It's mainly for the benefit of employees, as it provides a useful record of what happened in case you need time off work or need to claim compensation later on. But recording accidents also helps your employer to see what's going wrong and take action to stop accidents in future.

Sick pay

In most cases, if you need time off because of an accident at work, you'll only have the right to Statutory Sick Pay. Your employer may have a scheme for paying more for time off caused by accidents, or may decide to pay extra depending on what has happened.

Making an injury claim

If you've been injured in an accident at work and you think your employer is at fault, you may want to make a claim for compensation. Any claim must be made within three years of the date of the accident and you'll normally need a lawyer. If you belong to a trade union, you may be able to use their legal services. Or, you should speak to a specialist personal injury lawyer.

By law, your employer must be insured to cover a successful claim and should place a certificate with the name of their employer's insurance company where it can be seen at work. If not, they must give you the details if you need them.

If you're considering suing your employer, remember that the aim of legal damages is to put you in the position you would be in had the accident not happened - it's not about getting hold of some 'free' money. There are also court costs and legal fees to think about.

If you have an accident in the workplace, you should:

- make sure you record any injury in the 'accident book'
- if need be, make sure your employer has reported it to the HSENI
- check your contract or written statement of employment for information about sick or accident pay
- if there's a dispute, try to sort it out with your employer

PERSONAL PROTECTIVE EQUIPMENT (PPE) FOR INFECTION CONTROL

Personal Protective Equipment (PPE) is specialized clothing or equipment worn by an employee for protection against infectious materials.

PPE prevents contact with an infectious agent or body fluid that may contain an infectious agent, by creating a barrier between the potential i Importance

According to the hierarchy of controls by the National Institute for Occupational Safety and Health (NIOSH), PPE (sometimes also reffered to as PPE equipment)—is recommended to be the last level of defense to prevent occupational injuries, illnesses, and fatalities, but some businesses combined it with other control measures to ensure a safe and healthy environment for their workers. Here are some benefits of using PPEs:

- prevent unnecessary injury in the workplace;
- protect employees from excessive chemical exposure;
- prevent the spread of germs and infectious diseases including COVID-19;
- help businesses comply with regulatory requirements(e.g., The Personal Protective Equipment at Work Regulations 1992 that's recently been extended to limb workers); and
- Improve employee productivity and efficiency.

The 4 Basic Types

However, even the strictest controls will not necessarily eliminate all the risks associated with most job tasks and this is where the need for PPE must be evaluated. A hazard assessment can help identify which specialized PPE will be required. There are numerous types of workplace safety equipment available depending on the hazard exposure and work conditions. The following are basic PPE that can help protect employees:

Face and Eye Protection

PPE includes safety goggles and face shields and should be used for tasks that can cause eye damage or loss of vision, sprays of toxic liquids, splashes, and burns.

Safety Tips:

- Check if safety glasses comply with the ANSI Z87.1 eye protection standard.
- Ensure that there are no cracks or deformities on the lenses.
- Ensure the strap is in good working condition and is firmly sealed to the cheek and forehead.
- Clean and disinfect after use.

Respiratory Protection

PPE includes full-face respirators, self-contained breathing apparatus, gas masks, N95 respirators, and surgical masks are used for a task that can cause inhalation of harmful materials to enter the body. This includes harmful gas, chemicals, large-particle droplets, sprays, splashes, or splatter that may contain viruses and bacteria such as COVID-19, viral infections, and more.

Safety Tips:

- Ensure that the equipment is fit-tested and the employee has undergone proper training before wearing one.
- Carefully read the instructions to determine if it is designed to help protect against the hazards you may face.
- Change filters on half-mask or full-mask respirators frequently.

- Replace disposable respirators with every use.
- Surgical masks are not to be shared with anyone.
- Avoid touching the surgical mask after wearing it.
- Change surgical mask timely and should be disposed of after use.
- Replace the mask immediately if it is damaged or soiled.

Skin and Body Protection

PPE includes the following categories to protect employees from physical hazards:

Head Protection

PPE includes hard hats and headgears and should be required for tasks that can cause any force or object falling to the head.

Safety Tips:

- Ensure that there are no dents or deformities on the shell and connections are tightened inside.
- Do not store in direct sunlight as extreme heat can cause damage.
- Choose appropriate cleaning agents as it can weaken the shells of hard hats and may eliminate electrical resistance.
- Always replace a hard hat if it was used for any kind of impact, even if the damage is unnoticeable.

Body Protection

PPE includes safety vests and suits that can be used for tasks that can cause body injuries from extreme temperatures, flames and sparks, toxic chemicals, insect bites and radiation.

Safety Tips:

- Ensure that they are clean and free from cuts and burns.
- Always get a good fit to ensure full body protection.
- Ensure bodysuit is heat-resistant clothing when working with high-temperature hazards.

Hands Protection

PPE includes safety gloves and should be used for tasks that can cause hand and skin burns, absorption of harmful substances, cuts, fractures or amputations.

Safety Tips:

- Ensure hand protection fits perfectly with no spaces and is free from cuts, burns and chemical residue.
- Always replace them if any sign of contamination was observed.
- Use rubber gloves when working with heat and electricity to reduce the risk of burn or electrical shock.

Foot Protection

PPE includes knee pads and safety boots and should be used for tasks that can cause serious foot and leg injuries from falling or rolling objects, hot substances, electrical hazards, and slippery surfaces.

Safety Tips:

- Ensure boots have slip-resistant soles that can protect against compression and impact.
- Ensure the sole plate is in good condition to prevent punctures.

Fall Protection

PPE includes safety harnesses and lanyards and should be strictly used for tasks that can cause falling from heights and serious injury or death.

Safety Tips:

- Ensure that the straps are free from tears, deformities and burn marks.
- Check the buckles if connected securely and tightly.
- Dispose of the equipment if used after a falling incident.

Hearing Protection

PPE includes ear muffs and plugs and should be used for tasks that can cause hearing problems and loss of hearing.

Safety Tips:

- Ensure the equipment fit the ear canal perfectly.
- It is recommended to use formable earplugs to fit on different sizes of ear canals.
- Use protectors that reduce noise exposure to an acceptable level to have a room for communication.
- Ensure earplugs are clean and in good condition.

PPE Safety Requirements

To promote PPE safety in their workplace, safety officers will need to do the following:

- Check work sites regularly for the need of PPE.
- If PPE is needed, provide employees with properly-fitted PPE.
- Train employees on OSHA PPE standards.
- Provide protective goggles or face shields when there is a danger of flying particles or corrosive materials.
- Require that safety glasses are worn at all times in worksites that pose risk of eye punctures, abrasions, contusions, or burns.
- Provide and require protective gloves in situations where employees could be cut or be possibly exposed to corrosive liquids, chemicals, blood, and other potentially infectious materials.
- Require the use of foot protection when there is risk of foot injury from hot, corrosive, or poisonous substances, and falling objects.
- Inspect hard hats periodically for damage to the shell and suspension system.
- Maintain PPE in sanitary and ready-to-use conditions.
- Ensure that eyewash facilities and quick drench showers are easily accessible for employees when they are accidentally exposed to corrosive materials.
- Establish safe work procedures for disposing of or decontaminating PPE after hazardous exposuresSafetyCulture Marketplace: Your One-Stop Work Gear & Equipment Shop

Improve safety in your workplace with SafetyCulture Marketplace as your one-stop shop for all work gear and equipment needs. Get on-demand access to top quality and specialized work gear from trusted equipment brands in the industry—all in one centralized location. Raise the bar of safety and efficiency by empowering employees to request what they need with just a few taps, anytime and anywhere!

Ensure proper usage of PPE with PPE courses

Providing personal protective equipment (PPE) for your workers alone is not enough to protect them from hazards, injuries, and accidents, especially in industries such as construction, manufacturing, and healthcare. To ensure that their PPE fully serves its function, it's crucial that your workers understand its proper use, maintenance, and disposal to protect themselves (as well as the people they serve, in the case of healthcare workers). Thankfully, training is made easy and convenient for you with the available PPE courses online that your workers can take at their own convenience.

Here, you'll find our highly recommended PPE courses that will cover the different types of PPE and demonstrate when and how to use them properly. Reinforcing PPE training will not only keep your workers safe from hazards but also boost productivity without any threat of unwanted incidents. such as:

- Personal Protective Equipment (PPE) for Manufacturing
- Personal Protective Equipment (PPE) for Construction

PPE Safety Software

SafetyCulture (formerly iAuditor) provides a digital space for safety officers and employees to work towards PPE safety. SafetyCulture helps teams perform hazard assessments to determine if enforcing PPE use is the best control measure for the task. It also allows employees to maintain PPE with ease.

Using the SafetyCulture app, employees can take pictures of PPE defects and even annotate photos to show where the damage is. Tracking the number of usable PPE is easier for safety officers as well and they are better equipped to provide what their employees need to get the job done.

PPE Inspections with SafetyCulture

By using SafetyCulture for PPE inspections, safety officers are able to protect employees from the hazards of the job by ensuring that their PPEs are always in good condition.

With SafetyCulture , conducting regular PPE inspections is easy with convenient access to inspection data automatically stored in the cloud. Safety officers can also generate PPE inspection reports instantly with just one tap, anytime, and anywhere.

Beyond PPE inspections, SafetyCulture can also be used to develop an organization's health and safety program by enabling safety officers to do the following:

- Enforce standardized safety protocols across teams and different sites
- Mitigate risks with a wide variety of free risk assessment templates
- Plan corrective actions immediately to prevent incidents from occurring

Ensuring Safety and Protecting Employees

PPE Safety is the practice of ensuring a safe, working environment for employees and visitors through the use of Personal Protective Equipment (PPE). Safety is paramount to all businesses across industries. Using PPEs, paired with inspections such as workplace and restaurant inspections, assessments like health and safety risk assessments, and analysis such as gap analysis—is essential to protect employees from risks and hazards.

SWOT ANALYSIS

One way to explore these questions is through the use of another device partially borrowed from TQM: SWOT analysis. SWOT stands for Strengths, Weaknesses, Opportunities, and Threats. Each of the questions above can be examined in the light of SWOT analysis. What are the strengths and weaknesses of your organization in regard to each question? What opportunities exist for the organization in each area of its functioning? And what threats or challenges will the organization have to overcome if it is to continue to be successful, and to maintain quality performance?

Strengths and weaknesses: Strengths and weaknesses may be trends, rather than specifics. A level of service that is currently appropriate, for example, is not a strength if it is more or less than will be needed in a year. A new program that's not ready to implement yet is not a weakness if it's unready because the developers are taking the time necessary to make it effective.

By the same token, strengths and weaknesses don't necessarily lie only in the success of programs or the skills of staff members, but in such areas as relationships, contacts, and reputation. An organization running a great program may still be have serious weaknesses because it lacks some of these other features, no matter how well it carries out its day-to-day activities.

Opportunities: Opportunities can take many forms.

- An organization may be able to meet other needs with its current structure. For instance, an organization that publicizes and provides prenatal care to pregnant teens could be in an excellent position to also publicize and provide vaccinations, nutrition information, and help with parenting skills after the babies are born.
- It may be possible to expand into other areas of service, or into a larger arena (another town, another county, national instead of just one state).
- Increased funding may be available from new sources, or because of changed circumstances. A new census, for example, can result in an increase in federal funds to a region, or an economic downturn may bring a demand -- and increased funding -- for adult education or retraining.
- Collaboration with other groups, leading to increased resources, may become a prospect.
- Invitations or awards offered to your organization or staff members or good press may lead to your organization being viewed as more "legitimate."

Taking advantage of any opportunity can have both positive and negative consequences for your organization, so it's important to analyze the situation carefully before committing yourself.

Threats (Challenges): Some of the challenges that go along with any opportunity can be truly daunting if they're not thought through carefully. Many of the opportunities above require some sort of organizational restructuring or growth, processes that are always difficult, and require a lot of planning. Some even represent rethinking the purpose of the organization, which may become a different organization in the process. In becoming larger or more accepted, for instance, an organization may forget its roots or its guiding principles, and lose much of its effectiveness.

Other threats may come unaccompanied by opportunity. Your organization may experience difficulty finding -- and keeping -- ongoing funding and other resources, including competent staff; sustaining continued effort in all areas of functioning (advertising, recruitment, public relations, programming, evaluation, etc.); dealing with controversy; and addressing antagonism from individuals, other groups, or the community.

Applying SWOT analysis to all the areas your organization has to deal with makes it easier both to anticipate and prepare for the negative, and to remember to identify and build on the positiv

UNIT-8

LEGISLATION

a) Intellectual Property Rights(IPR), Patents, Trademarks, Copyrights
b) Features of Factories Act 1948 with Amendment (only salient points)
c) Features of Payment of Wages Act 1936 (only salient points)

INTELLECTUAL PROPERTY RIGHTS (IP RIGHTS)

Intellectual Property Rights (IP Rights) are like any other property rights which are intangible in nature. The IP Rights usually give the creator an exclusive right over the use of his/her creation for a certain period of time. With the rapid increase in the globalization and opening up of the new vistas in India, the "Intellectual Capital" has become one of the key wealth drivers in the present era. There are different country specific legislations, as well international laws and treaties that govern IP rights.

Every startup has IP Rights, which it needs to understand and protect for excelling in its business. Every startup uses trade name, brand, logo, advertisements, inventions, designs, products, or a website, in which it possesses valuable IP Rights. While starting any venture, the startup also needs to confirm that it is not in violation of the IP Rights of any other person to save itself from unwarranted litigation or legal action which can thwart its business activities. Further, startup ventures should be proactive in developing and protecting their intellectual property for many reasons like improving the valuation of its business, to generate better goodwill, to protect its competitive advantage, to use intellectual property as a marketing edge and to use the IP Rights as a potential revenue stream through licensing.

IP Rights protect several aspects of a business and each type of IP Right carries its own advantages. The scope of IP Rights is very wide, but the prime areas of intellectual property which are of utmost importance for any startup venture are as follows:

- **Trademarks**
- **Patents**
- **Copyrights and Related Rights**
- **Industrial Designs**
- **Trade Secrets**

TRADEMARKS

The Trade Marks Act 1999 ("**TM Act**") provides, inter alia, for registration of marks, filing of multiclass applications, the renewable term of registration of a trademark as ten years as well as recognition of the concept of well-known marks, etc. It is pertinent to note that the letter "R" in a circle i.e. ® with a trademark can only be used after the registration of the trademark under the TM Act.

Trademarks means any words, symbols, logos, slogans, product packaging or design that identify the goods or services from a particular source. As per the definition provided under Section 2 (zb) of the TM Act, "trade mark" means a mark capable of being represented graphically and which is capable of distinguishing the goods or services of one person from those of others and may include shape of goods, their packaging and combination of colors.

The definition of the trademark provided under the TM Act is wide enough to include non-conventional marks like color marks, sound marks, etc. As per the definition provided under Section 2 (m) of the TM Act, "mark" includes a device, brand, heading, label, ticket, name, signature, word, letter, numeral, shape of goods, packaging or combination of colors or any combination thereof.

Accordingly, any mark used by the startup in the trade or business in any form, for distinguishing itself from other, can qualify as trademark. It is quite significant to note that the Indian judiciary has been proactive in the protection of trademarks, and it has extended the protection under the trademarks law to Domain Names as demonstrated in landmark cases of Tata Sons Ltd v Manu Kosuri & Ors [90 (2001) DLT 659] and Yahoo Inc. v Akash Arora [1999 PTC 201].

Points to Consider While Adopting a Trademark

Any startup needs to be cautious in selecting its trade name, brands, logos, packaging for products, domain names and any other mark which it proposes to use. You must do a proper due diligence before adopting a trademark. The trademarks, can be broadly classified into following five categories:

a. **Generic**
b. **Descriptive**
c. **Suggestive**
d. **Arbitrary**
e. **Invented/Coined**

Generic marks means using the name of the product for the product, like "Salt" for salt.

Descriptive marks means the mark describing the characteristic of the products, like using the mark "Fair" for the fairness creams.

Suggestive marks means the mark suggesting the characteristic of the products, like "Habitat" for home furnishings products.

Arbitrary marks means mark which exist in popular vocabulary, but have no logical relationship to the goods or services for which they are used, like "Blackberry" for phones.

The invented/ coined marks means coining a new word which has no dictionary meaning, like "Adidas". The strongest marks, and thus the easiest to protect, are invented or arbitrary marks. The weaker marks are descriptive or suggestive marks which are very hard to protect. The weakest marks are generic marks which can never function as trademarks.

India follows the NICE Classification of Goods and Services for the purpose of registration of trademarks. The NICE Classification groups products into 45 classes (classes 1-34 include goods and classes 35-45 include services). The NICE Classification is recognized in majority of the countries and makes applying for trademarks internationally a streamlined process. Every startup, seeking to trademark a good or service, has to choose from the appropriate classes, out of the 45 classes.

While adopting any mark, the startup should also keep in mind and ensure that the mark is not being used by any other person in India or abroad, especially if the mark is well-known. It is important to note that India recognizes the concept of the "Well-known Trademark" and the principle of "Trans-border Reputation".

Example of well-known trademarks are Google, Tata, Yahoo, Pepsi, Reliance, etc. Further, under the principle of "Trans-border Reputation", India has afforded protection to trademarks like Apple, Gillette, Whirlpool, Volvo, which despite having no physical presence in India, are protected on the basis of their trans-border reputation in India.

Enforcement of Trademark Rights

Trademarks can be protected under the statutory law, i.e., under the TM Act and the common law, i.e., under the remedy of passing off. If a person is using a similar mark for similar or related goods or services or is using a well-known mark, the other person can file a suit against that person for violation of the IP rights irrespective of the fact that the trademark is registered or not.

Registration of a trademark is not a pre-requisite in order to sustain a civil or criminal action against violation of trademarks in India. The prior adoption and use of the trademark is of utmost importance under trademark laws.

The relief which a court may usually grant in a suit for infringement or passing off includes permanent and interim injunction, damages or account of profits, delivery of the infringing goods for destruction and cost of the legal proceedings. It is pertinent to note that infringement of a trademark is also a cognizable offence and criminal proceedings can also be initiated against the infringers.

PATENTS

Patent, in general parlance means, a monopoly given to the inventor on his invention to commercial use and exploit that invention in the market, to the exclusion of other, for a certain period. As per Section 2(1) (j) of the Patents Act, 1970, "invention" includes any new and useful;

i. **art, process, method or manner of manufacture;**
ii. **machine, apparatus or other article;**
iii. **substance produced by manufacture, and includes any new and useful improvement of any of them, and an alleged invention;**

The definition of the word "Invention" in the Patents Act, 1970 includes the new product as well as new process. Therefore, a patent can be applied for the "Product" as well as "Process" which is **new**, involving inventive step and capable of industrial application can be patented in India.

The invention will not be considered new if it has been disclosed to the public in India or anywhere else in the world by a written or oral description or by use or in any other way before the filing date of the patent application. The information appearing in magazines, technical journals, books etc, will also constitute the prior knowledge. If the invention is already a part of the state of the art, a patent cannot be granted. Examples of such disclosure are displaying of products in exhibitions, trade fairs, etc. explaining its working, and similar disclosures in an article or a publication.

It is important to note that any invention which falls into the following categories, is not patentable: (a) frivolous, (b) obvious, (c) contrary to well established natural laws, (d) contrary to law, (e) morality, (f) injurious to public health, (g) a mere discovery of a scientific principle, (h) the formulation of an abstract theory, (i) a mere discovery of any new property or new use for a known substance or process, machine or apparatus, (j) a substance obtained by a mere admixture resulting only in the aggregation of the properties of the components thereof or a process for producing such substance, (k) a mere arrangement or rearrangement or duplication of known devices, (l) a method of agriculture or horticulture, and (m) inventions relating to atomic energy or the inventions which are known or used by any other person, or used or sold to any person in India or outside India. The application for the grant of patent can be made by either the inventor or by the assignee or legal representative of the inventor. In India, the term of the patent is for 20 years. The patent is renewed every year from the date of patent.

Use of Technology or Invention

While using any technology or invention, the startup should check and confirm that it does not violate any patent right of the patentee. If the startup desires to use any patented invention or technology, the startup is required to obtain a license from the patentee.

Enforcement of Patent Rights

It is pertinent to note that the patent infringement proceedings can only be initiated after grant of patent in India but may include a claim retrospectively from the date of publication of the application for grant of the patent. Infringement of a patent consists of the unauthorized making, importing, using, offering for sale or selling any patented invention within the India. Under the (Indian) Patents Act, 1970 only a civil action can be initiated in a Court of Law. Like trademarks, the relief which a court may usually grant in a suit for infringement of patent includes permanent and interim injunction, damages or account of profits, delivery of the infringing goods for destruction and cost of the legal proceedings.

COPYRIGHT

Copyright means a legal right of an author/artist/originator to commercially exploit his original work which has been expressed in a tangible form and prevents such work from being copied or reproduced without his/their consent.

Under the Copyright Act, 1957, the term "work", in which copyright subsists, includes an artistic work comprising a painting, a sculpture, a drawing (including a diagram, a map, a chart or plan), an engraving, a photograph, a work of architecture or artistic craftsmanship, dramatic work (recitation, choreographic work), literary work (including computer programmes, tables, compilations and computer databases), musical work (including music as well as graphical notations), sound recording and cinematographic film.

In the case of original literary, dramatic, musical and artistic works, the duration of copyright is the lifetime of the author or artist, and 60 years counted from the year following the death of the author and in the case of cinematograph films, sound recordings, posthumous publications, anonymous and pseudonymous publications, works of government and works of international organizations are protected for a period of 60 years which is counted from the year following the date of first publication.

In order to keep pace with the global requirement of harmonization, the Copyright Act, 1957 has brought the copyright law in India in line with the developments in the information technology industry, whether it is in the field of satellite broadcasting or computer software or digital technology.

Registration of Copyright

In India, the registration of copyright is not mandatory as the registration is treated as mere recordal of a fact. The registration does not create or confer any new right and is not a prerequisite for initiating action against infringement. The view has been upheld by the Indian courts in a catena of judgments. Despite the fact that the registration of copyright is not mandatory in India and is protectable through the International Copyright Order, 1999, it is advisable to register the copyright as the copyright registration certificate is accepted as a "proof of ownership" in courts and by police authorities, and acted upon smoothly by them.

Enforcement of Copyright in India

Any person who uses the original work of the other person without obtaining license from the owner, infringes the copyright of the owner. The law of copyright in India not only provides for civil remedies in the form of permanent injunction, damages or accounts of profits, delivery of the infringing material for destruction and cost of the legal proceedings, etc, but also makes instances of infringement of copyright, a cognizable offence punishable with imprisonment for a term which shall not be less than six months but which may extend to three years, with a fine which shall not be less than INR 50,000 but may extend to INR 200,000

For the second and subsequent offences, there are provisions for enhanced fine and punishment under the Copyright Act. The (Indian) Copyright Act, 1957 gives power to the police authorities to register the Complaint (First Information Report, i.e., FIR) and act on its own to arrest the accused, search the premises of the accused and seize the infringing material without any intervention of the court.

INDUSTRIAL DESIGNS

As per the definition given under Section 2(d) of the Designs Act, 2000, "design" means only the features of shape configuration patterns or ornament applied to any article by any industrial process or means whether manual mechanical or chemical separate or combined which in the finished article appeal to and are judged solely by the eye. However, "design" does not include any mode or principle of construction or anything which is in substance a mere mechanical device and does not include any trademark as defined under the TM Act or any artistic work as defined under the Copyright Act, 1957. The total period of validity of registration of an Industrial Design under the (Indian) Designs Act, 2000 is 15 years.

Features of shape, configuration, pattern, ornament or composition of lines or colours applied to any article, whether in two dimensional or three dimensional or in both forms, can be registered under the (Indian) Designs Act, 2000. However, functionality aspects of a design are not protected under the (Indian) Designs Act, 2000, as the same are subject matter of patents.

Design of an article is not registrable in India, if it:

- is not new or original;
- has been disclosed to the public anywhere in India or in any other country by publication in tangible form or by use in any other way prior to the filing date or priority date of the application;
- is not significantly distinguishable from known designs or combination of known designs; or
- comprises or contains scandalous or obscene matter.

Enforcement of Design Rights in India

The (Indian) Designs Act, 2000, only provides for civil remedies. Besides injunction, monetary compensation is recoverable by the proprietor of the design either as contract debt or damages. An action for infringement of design can only be initiated after the registration of the design, however, an action for passing-off is maintainable in case of unregistered design.

TRADE SECRETS

Trade secrets includes any confidential business information which provides an enterprise a competitive edge over others. Trade secrets encompass manufacturing or industrial secrets and commercial secrets, formula, practice, process, design, instrument, pattern, commercial method, or compilation of information which is not generally known or reasonably ascertainable by other.

The unauthorized use of such information by persons other than the holder is regarded as an unfair practice and a violation of the trade secret. There are no specific statutes under the Indian law for the protection of trade secrets and the same are protectable under the common law rights.

STRATEGIES FOR PROTECTION AND EXPLOITATION OF IPR FOR STARTUPS

1. **Make Intellectual Property protection a priority:**

   Start-ups cannot afford the complete protection available under the intellectual property regime. The first step for any startup is to evaluate and prioritize the IP Rights involved in its business. Depending upon the type of industry involved, IP Rights play an important role. Failure to identify or prioritize IP Rights, is likely to create problems for startup's business, especially during negotiations with future investors or exiting its business. Sometimes IP Rights are the only asset available with a startup.
2. **Register Intellectual Property Rights:**

   It is important to note that certain IP Rights like patents and designs are required to be registered before claiming any protection under the respective statutes. On the other hand, certain IP Rights like trademark and copyright need not be mandatorily registered for protection under. Nevertheless, a registered IP Right carries a greater value and acts as evidence of use of the IP Rights before courts as well as enforcement agencies;

3. **Due Diligence of IP Rights:**

   For any startup, it is indispensable that it does not violate IP Rights of any other person. This will ensure safety from unwarranted litigation or legal action which can thwart its business activities. This makes it even more important for startups to make careful IP decisions in the initial phase and conduct proper due diligence of IP Rights, which it is using or intends to use.

4. **Implement clear and effective policies and strategies for protection of IP Rights:**

   It is in the long term interest of startups to have an Intellectual Property Policy for management of various IP rights which may be presently owned, created or acquired in future by startups. The aim of such a policy is to ensure that there are no inter-se dispute between the promoters of the startups, which remains till date to be one of the main concerns for failure of startups.

5. **Agreements related to Intellectual Property:**

   It is pertinent to note that having proper documentation in the form of agreements like non-disclosure agreements, agreements with employees or independent contractors, can make all the difference between the success and failure of startups. Usually, intellectual property is created either by the founders or some key employee or a third party. The intellectual property so created, must be protected through a proper agreement between the founder or key employee or a third party, as the case may be and the startup. If the agreement, with founders or employees or a third party, , under which a novel idea was/is created, is overlooked, it could create bottlenecks later after such idea becomes successful. Accordingly, the startups need to ensure that anything created on behalf of the startup, belongs to the startup and not the Employee or a third party. Further, it is advisable to enter into elaborate assignments, licensing or user agreements, and care should be taken to make provisions for all post termination IP Right issues.

What is intellectual property and what is intellectual property rights?

The World Intellectual Property Organisation defines intellectual property as creations of the mind, such as inventions; literary and artistic works; designs; and symbols, names and images used in commerce.

In simple terms, anything that you create by applying your mind or intellect becomes your intellectual property. For example, you are setting up a business of your own and have designed a logo for your brand. That logo is your intellectual property. Similarly a scientific invention is also an intellectual property.

When you create a logo for your brand, you will want it to be unique and resonate with your brand. You want to stand out from your competitors and also make sure your ideas are not copied or misused. This is where intellectual property rights come in.

Intellectual property rights are given to creators for a specific amount of time. During this time the creator is protected from infringement and can license their product for manufacturing in return for a royalty.

The legal rights provided to the creator or inventor of intellectual property are known as intellectual property rights. Every creator enjoys certain privileges provided by the law that protects the creators from infringement or misuse of their work without their explicit permission. Intellectual property rights also aid creators in putting their work to use for the general public by allowing for commercial licencing.

Intellectual property rights are granted to creators for a limited period of time.

What are the types of intellectual property?

As mentioned earlier, intellectual property is an umbrella term used to describe creations of the human mind. For the purpose of intellectual property rights protection, IP can be categorised as the following:

1. Trademarks: This refers to any symbols, logos signs, words and/ or sounds that are unique to a product or business. For example, the trademark blue bird logo for Twitter is its representative of its brand and is protected by intellectual property rights.
2. Copyrights: This is applicable to any published work such as books, art work, songs, web content and films, among others. Your favourite cartoon duo Tom and Jerry are artwork copyrighted by Disney.
3. Patents: Any new invention that holds the potential being put to commercial use is protected from infringement or duplication by a patent. Recently, pharma major Cipla lost a patent battle to Amgen Inc over the injectable cancer drug Kyprolis.
4. Industrial designs: These are features such as shape, pattern, ornament or composition applied to a product by an industrial process such as by chemical process or mechanical process.
5. Geographical indicators: A name or sign used on certain products that marks the identity of a particular geographical location or origin is termed as a geographical indicator. For example, West Bengal holds the geographical indicator tag for its beloved Rosogolla.

Indian laws to protect intellectual property

In India, intellectual property rights are protected by a host of laws and rules issued by the government from time to time. Primarily, intellectual property rights are protected by the following legislation and rules:

1. Trade Marks Act, 1999 and Trade Marks Rules of 2002 and 2017. Your beloved mango drink Maaza was embroiled in a trademark controversy when Coca Cola claimed infringement of trademark over Bisleri. Originally a drink by Bisleri, the trademarks for the drink was sold to Coca Cola by Bisleri. However, in 2008, Bisleri filed for registering the trademark for Maaza in Turkey, which was then contested by Coca Cola
2. The Patents Act, 1970 (as amended in 2005), 2003 Patent Rules and 2016 Patent Amendment Rules
3. The Copyright Act, 1957. The most notable case under this Act was the Tips vs Wynk music fallout. Tips Industries and streaming platform Wynk had entered an agreement where the latter was granted a licence to access the large music database of Tips. At the end of this agreement, amid negotiations for extending the agreement, Tips alleged that Wynk was infringeing on its IP. The judgement was in favour of Tips Industries.
4. The Designs Act, 2000, and Designs Rules, 2001
5. The Geographical Indications of Goods (Registration and Protection) Act, 1999
6. The Semiconductor Integrated Circuits Layout Design Act, 2000
7. The Protection of Plant Varieties and Farmers' Right Act, 2001

While these laws provide domestic protection to intellectual property rights, India has also signed agreements with various countries for safeguarding intellectual property.

How to protect your intellectual property in India

The intellectual property laws in the country are designed to support innovation and research. In order to reap the benefits of intellectual property rights in the country, the first step is to register your creation. You can register your intellectual property for patents, trademarks or copyrights. In case of an infringement, you can challenge the same as civil disputes in the court of law.

Advantages of intellectual property rights

Here's why it is important to look after intellectual property rights:

1. **Ownership**: In a competitive world, entrepreneurs, businesses and researchers are constantly trying to outdo each other. A determining factor to win in this competition is ownership of the creation. Intellectual property rights ensure that the creator holds the sole authority over their creation.
2. **Infringement**: With intellectual property rights, creators are protected against any competitor illegally contravening on the creation. For example, an entrepreneur's scientific invention is their own. They are the sole owner of the creation and can choose to do what they wish with it. Anybody using the scientific invention for any purpose without paying due credit or royalty to the creator is liable to face legal action.
3. **Commercialisation**: Intellectual property rights are of great importance when a creator wishes to commercialise their product. It paves the way for the creator to issue a licence for mass production of the product in exchange for monetary benefits and royalty.
4. **Learning**: While registering a new intellectual property, the creator has to reveal details of the product to the general public. This practice encourages a culture of sharing of information that can be used for further development by contemporaries in the field.

Final note: Intellectual property rights are important drivers of the economic growth of a country. For inventors and entrepreneurs, who are constantly competing with each other to stand out in the market, intellectual property rights are essential.

In India, intellectual property rights safeguard the interests of the inventor. You can bring civil or criminal cases against competitors to enforce your intellectual property rights, depending on the infringement or violation.

There are only three ways to protect intellectual property in the United States: through the use patents, trademarks or copyrights. A patent applies to a specific product design; a trademark to a name, phrase or symbol; and a copyright to a written document. All three methods have limitations--there's no one perfect way to protect an idea.

Copyright Protection

A copyright will protect the following categories of works:

- literary works
- musical works, including any accompanying words
- dramatic works, including any accompanying music
- pantomimes and choreographic works
- pictorial, graphic and sculptural works
- motion pictures and other audiovisual works
- sound recordings
- architectural works
- computer programs (sometimes the graphical user interface) and websites

Copyright protection gives the copyright holder the exclusive right to copy the work, modify it (that is, create "derivative works"), and distribute, perform and display the work publicly.

Ideas or concepts do not have copyright protection. Copyright protects the expression of the idea, but not the ideas themselves. For example, if I ask you what a chair is, you get a picture in your head; the picture I get in my head is different from the picture you get in your head and probably also different from the picture Buffy gets in her head. These are the "ideas" of what a chair is. However, if you were to draw the chair you envisioned in your head or use words to describe that chair, it's an "expression" of the idea--and that's what's protected by copyright.

Generally, the only protection for ideas and concepts is through trade secret law and/or confidentiality agreements, which provide a contractual remedy for misuse or disclosure of the idea.

Patents

Patents protect processes, methods and inventions that are "novel," "non-obvious" and "useful." If granted, a patent gives you a 20-year monopoly on selling, using, making or importing an invention into the United States. The requirements for a patent are complex, but here they are in a nutshell:

- **Your work must be novel.** This means it must not be known or used by others in this country, or patented or described in a printed publication here or abroad, or in public use or for sale in this country more than one year prior to the application for patent.
- **Your work must be non-obvious.** This means it must not be obvious to a person having ordinary skill in the pertinent art as it existed when the invention was made.
- **Your work must be useful.** This means that it must have current, significant, beneficial use as process, machine, manufacture, composition of matter or improvements to one of these. According to the Patent Office: "The word 'process' is defined by law as a process, act or method, and primarily includes industrial or technical processes. The term 'machine' used in the statute needs no explanation. The term 'manufacture' refers to articles that are made, and includes all manufactured articles. The term 'composition of matter' relates to chemical compositions and may include mixtures of ingredients as well as new chemical compounds. These classes of subject matter taken together include practically everything which is made by man and the processes for making the products."

Patent protection requires full public disclosure of the work in detail and therefore precludes maintaining any trade secret protection in the same work.

Trademarks

A trademark is like a brand name. It is any word(s) or symbol(s) that represent a product to identify and distinguish it from other products in the marketplace. A trademark word example would be "Rollerblades." A trademark symbol would be the peacock used by NBC.

A trademark can be registered in three ways:

1. By filing a "use" application after the mark has been used.
2. By filing an "intent to use" application if the mark has not yet been used.
3. In certain circumstances in which a foreign application exists, you can rely on that.

The (TM) mark may be used immediately next to your mark. The ® registration symbol may only be used when the mark is registered with the PTO. It is unlawful to use this symbol with your mark before receiving an issued registration from the PTO.

What qualities make for a strong trademark? The cardinal rule is that a mark must be distinctive. The more distinctive it is, the easier your trademark will be to enforce. This is why so many trademarked products have unique spellings.

Trademark rights last indefinitely if the company continues to use the mark to identify its goods or services. When the mark is no longer being used, the registration is terminated. The initial term of federal trademark registration is 10 years, with 10-year renewal terms.

Trade Secrets

There is a great deal of confusion regarding trade secrets. Many people think that a trade secret is some type of protection provided by the government that allows them to seek recourse in court should someone infringe upon their idea. However, unlike copyrights, trademarks and patents, a trade secret is not registered with any government office to provide a verifiable public record of any claims to the secret. You can, however, declare one to a patent lawyer in a notarized and signed disclosure. In this manner the trade secret belongs to you forever--or until someone leaks it.

Trade secrets refer to items such as recipes that are unique and provide a business with a competitive advantage, but which cannot be safeguarded under current forms of idea protection such as copyright, trademark or patent. The best form of protection for these items is to keep them a secret. One of the most famous and best-kept trade secrets is the formula for Coca-Cola.

The best way to secure the information for a trade secret is to restrict access to the secret and have individuals and companies sign nondisclosure agreements with you should you enter into a relationship with them which will require them to know some aspects of the secret. If someone independently develops or reverse-engineers your trade secret, there's nothing you can do. If someone does leak it, you can sue for theft. Suing, however, cannot stop the person from using the leaked information. So although you may get money from the suit, you lose the larger potential profits you could have made from the idea. Still, if your luck holds and your trade secret remains secret, royalty income from it can last significantly longer than the patent period.

FEATURES OF FACTORY ACT 1948

Introduction

The Factories Act of 1948 was enacted to protect the welfare of workers in a factory by regulating employment conditions, working conditions, the working environment, and other welfare requirements of specific industries. The Court held in Ravi Shankar Sharma v. State of Rajasthan (1993) that the Factory Act is social legislation that covers the health, safety, welfare, and other aspects of factory workers. The Factories Act lays out guidelines and safety measures for using machinery, and with its strict compliance, it also provides owners with instructions. When factory workers were taken advantage of and exploited by paying them low wages, the Factories Act was passed.

A factory is a building or group of buildings where people work with machinery to make goods. The primary goal of the Factories Act is to safeguard employees in a factory from industrial and occupational risks. This Act gives the owner or occupier of a factory a particular responsibility to secure and protect employees from employment in conditions harmful to their health and safety in order to safeguard workers. It is stated in the Act that the purpose of the Factories Act is to amend and consolidate the legal framework governing factory labour. The Bhopal gas tragedy case (1984) raised public awareness of factory pollution and risks, necessitating government action to allow legislation amendments.

Salient features of Factories Act, 1948 with Amendment-

The important features of the 1948 Act are as follows:

- The word "factory" has been expanded by the Factories (Amendment) Act of 1976 to include contract labour when determining whether a factory has a maximum of 10 or 20 employees.
- The Act increased the minimum age for children to work in workplaces from 12 to 14 and reduced their daily working hours from 5 to 4 and a half.
- The Act forbids women and children from working in factories from 7 p.m. to 6 a.m.
- The difference between a seasonal and non-seasonal factory has been abolished by the Act.
- The Act, which has provisions for factory registration and licencing.
- The state government is required to make sure that all factories are registered and also have valid licences that are renewed from time to time.
- The Act gives state governments the authority to enact rules and regulations that ask for management and employee association for the benefit of employees.
- The state government has the authority to apply the Act's requirements to any establishment, regardless of the number of employees inside and regardless of whether the establishment engages in manufacturing operations.
- In Rabindra Agarwal v. State of Jharkhand (2010), the Jharkhand High Court held that the Factories Act, special legislation would prevail over the Indian Penal Code

Objectives of Factories Act, 1948

- The major goal of the Factories Act of 1948 is to establish adequate safety measures and to enhance the health and welfare of workers employed in a factory. The Act also protects workers from various industrial and occupational hazards.
  - **Heath:** According to the Act, all factories must be kept clean, and all essential safeguards must be taken to safeguard the health of workers. The factory must have a sufficient drainage system, adequate lighting, ventilation, temperature, etc. There must be clean water supplies. Separate restrooms and urinals must be built in convenient locations for males and females. These must be freely accessible to employees and kept clean.
  - **Safety:** The Act requires that machines be properly fenced; that no young adults work on any dangerous machines in enclosed places, and also that appropriate manholes be provided so that employees may escape in an emergency.
  - **Welfare:** The Act specifies that appropriate and suitable washing facilities for workers must be provided and maintained in every factory. There must be storage and drying facilities, as well as sitting areas, first-aid equipment, shelters, restrooms and lunch rooms.
- The Act also imposes some restrictions on the employment of women, small children, and teenagers, such as working hours, intervals, holidays, etc., as well as on annual leave with pay, etc.
  - **Working hours:** The Act sets working hours for all workers, and no adult worker must be permitted to work in a workplace for more than 48 hours per week. Weekly holidays need to be granted.
- The Act also imposes specific restrictions on owners, occupiers, or the manufacturer's head in order to safeguard employees and ensure their health and safety precautions.
- The Act protects workers from exploitation and improves working conditions and the environment within factory premises.
  - **Penalties:** The Act also specifies specific rules created with provisions under the Act, and written orders that are violated. It is an offence, and penalties will be imposed, imprisonment for up to a year; a fine of up to one lakh rupees; or both fine and

imprisonment. Any employee who misuses equipment related to the welfare, safety, and health of other employees, or those connected to the performance of his duties, suffers a Rs.500 fine.

What is "Wages"?

Wage is monetary compensation or remuneration, personnel expenses, of the labor paid by an employer to the employees in exchange for work done. Wage may be calculated as a fixed amount for each task completed or at an hourly / daily rate / based on an easily measured quantity of work done.

Wages include all remunerations which are expressed in terms of money and **"includes"** the following:

- Amount payable under the terms of employment
- Payable under any award / settlement or order of a Court
- Paid as Overtime work or for holidays / leave period
- Payable on account of Termination of employment"

Wages does **"not include"** the following Payments:

- Bonus which does not form a part of remuneration
- Value of any house accommodation, supply of light, water, medical assistance etc.
- Any traveling concession
- Contribution payable by employer to any pension or provident fund
- Any sum paid to defray special expenses incurred as requirement of the Employment
- Gratuity payable on termination of employment

Objectives:

- Regulate payment of wages to specific class of workers employed in industry without any wrongful deductions apart from what is mentioned in the Act
- Define regulations around fixing of wage period, time and mode of payment of wages
- Regulates the rights of the workers covered under this Act

Applicability:

This Act is applicable to all persons employed, whether directly or through contractors, in a factory or certain specified industrial or other establishments.

- The Central Government is responsible for enforcement of the Act in Railways, Mines, Oilfields and air transport services.
- The State Government are responsible for it all other establishment (factories and other establishments)
- The Act does not cover those whose wage is Rs. 24,000/- or more per month

Salient Features of the Act:

A. **Obligations of Employers:**

Every employer is responsible for the payment of wages to all the employees that he employs. Additionally, apart from the Employer, all the person so named / person so responsible to the employer / the person so nominated shall also be responsible for such payment.

B. **Wage Period:**

Every person responsible for Wage Payment shall fix periods in respect of which such wages shall be payable. No wage-period shall exceed one month.

C. **Time and Mode of Payment of Wages:**

Every establishment having employees in excess of 1,000 person shall pay the wages before the expiry of the 10^{th} day. All other employers shall make the Wage payment by the expiry of 07^{th} day.
Employers shall make the payment of wages in current currency notes i.e. cash or via Bank transfer.

D. **Deductions from Wages:**

Employers shall ensure that wages are paid to all employees without deduction of any kind except those authorized by or under this Act.
Deduction includes the reduction of wages for the following:

- Fines
- Absence from duty
- Damage to or loss of goods including loss of money where such damage or loss is directly attributable to employee's neglect or default
- Recovery of advances / loans and the interest due in respect thereof
- Adjustment of over-payments of wages
- Payments made by the employed person to the employer or his agent shall also be deemed to be a deduction from wages.

Deductions **does not include** following penalties (provided the rules w.r.t. penalty are in conformity with the requirements as specified by the State Government):

- Withholding of increment or promotion
- Reduction to a lower post
- Suspension

A. **Compliance Requirements:**

Maintenance of Registers:
Employers shall maintain Registers and records giving the following details:

- Particulars of persons employed
- Work performed by Employees
- Wages paid to employees and deductions made from their wages

All these Registers are required to be preserved for a period of 03 Years.

Display of Notice of Abstracts:

All Employers shall display Notice containing such abstracts of this Act and of the rules made thereunder in English and in the language of the majority of the persons employed in the factory. It includes the following:

- List of acts and omissions approved under Rule 12
- Rates of wages payable to employees (excluding those in supervisor / manager positions).

A. Penalties for Offences under the Act:

If Employers fails to maintain registers required under this act or wilfully refuses to furnish such information or return or wilfully furnishes false information shall be punishable with fine up to Rs. 1,500 and may extend to Rs. 7,500.

Additionally, If Employers fails to pay the wages by the date fixed by the authority shall be punishable with an additional fine which may extend to Rs. 750 / Day.

B. Un-disbursed Wages in event of Death of an Employee:

In case the Wages cannot be paid on account death of an employee or non-availability of details of his / her whereabouts, such amounts is required to be paid to person nominated by him in this behalf. In case employee has not nominated anyone, such amount shall be deposited with the prescribed authority.

UNIT-9

SMART TECHNOLOGY

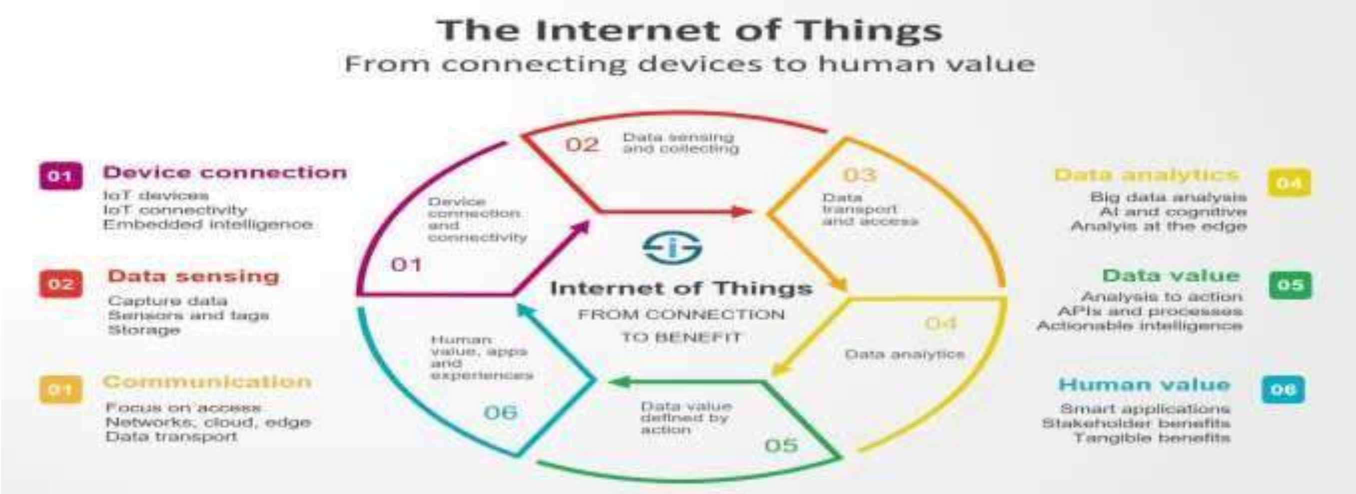
- **Concept of IOT, How IOT works**
- **Components of IOT, Characteristics of IOT, Categories of IOT**
- **Applications of IOT- Smart Cities, Smart Transportation, Smart Home, Smart Healthcare, Smart Industry, Smart Agriculture, Smart Energy** Management etc.

The Internet of Things (IoT) describes the network of physical objects—"things"—that are embedded with sensors, software, and other technologies for the purpose of connecting and exchanging data with other devices and systems over the internet.

"The Internet of Things (IoT) is a system of interrelated computing devices, mechanical and digital machines, objects, animals or people that are provided with unique identifiers and the ability to transfer data over a network without requiring human-to-human or human-to-computer interaction."

Internet of Things :

According to the definition of IoT, It is the way to interconnection with the help of the internet devices that can be embedded to implement the functionality in everyday objects by enabling them to send and receive data. Today data is everything and everywhere. Hence, IoT can also be defined as the analysis of the data generate a meaning action, triggered subsequently after the interchange of data. IoT can be used to build applications for agriculture, assets tracking, energy sector, safety and security sector, defense, embedded applications, education, waste management, healthcare product, telemedicine, smart city applications, etc.



However, all complete IoT systems are the same in that they represent the integration of four distinct components: sensors/devices, connectivity, data processing, and a user interface.

How an IoT System Actually Works

As you saw in the previous chapter, the applications for IoT extend across a broad variety of use cases and verticals. However, all complete IoT systems are the same in that they represent the integration of four distinct components: sensors/devices, connectivity, data processing, and a user interface.

We'll outline what each one means in the sections below and how they come together to form a complete IoT system. Each of these sections will also serve as the organizational structure of the rest of this ebook, and we'll dive into these components more deeply in the chapters to follow.

1) Sensors/Devices

First, sensors or devices collect data from their environment. This data could be as simple as a temperature reading or as complex as a full video feed.

We use "sensors/devices," because multiple sensors can be bundled together or sensors can be part of a device that does more than just sense things. For example, your phone is a device that has multiple sensors (camera, accelerometer, GPS, etc), but your phone is not just a sensor since it can also perform many actions.

However, whether it's a standalone sensor or a full device, in this first step data is being collected from the environment by *something*.

2) Connectivity

Next, that data is sent to the cloud, but it needs a way to get there!

The sensors/devices can be connected to the cloud through a variety of methods including: cellular, satellite, WiFi, Bluetooth, low-power wide-area networks (LPWAN), connecting via a gateway/router or connecting directly to the internet via ethernet (don't worry, we'll explain more about what these all mean in our connectivity section).

Each option has tradeoffs between power consumption, range, and bandwidth. Choosing which connectivity option is best comes down to the specific IoT application, but they all accomplish the same task: getting data to the cloud.

3) Data Processing

Once the data gets to the cloud (we'll cover what the cloud means in our data processing section)), software performs some kind of processing on it.

This could be very simple, such as checking that the temperature reading is within an acceptable range. Or it could also be very complex, such as using computer vision on video to identify objects (such as intruders on a property).

But what happens when the temperature is too high or if there is an intruder on property? That's where the user comes in.

4) User Interface

Next, the information is made useful to the end-user in some way. This could be via an alert to the user (email, text, notification, etc). For example, a text alert when the temperature is too high in the company's cold storage.

A user might have an interface that allows them to proactively check in on the system. For example, a user might want to check the video feeds on various properties via a phone app or a web browser.

However, it's not always a one-way street. Depending on the IoT application, the user may also be able to perform an action and affect the system. For example, the user might remotely adjust the temperature in the cold storage via an app on their phone.

And some actions are performed automatically. Rather than waiting for you to adjust the temperature, the system could do it automatically via predefined rules. Rather than just call you to alert you of an intruder, the IoT system could also automatically notify security teams or relevant authorities.

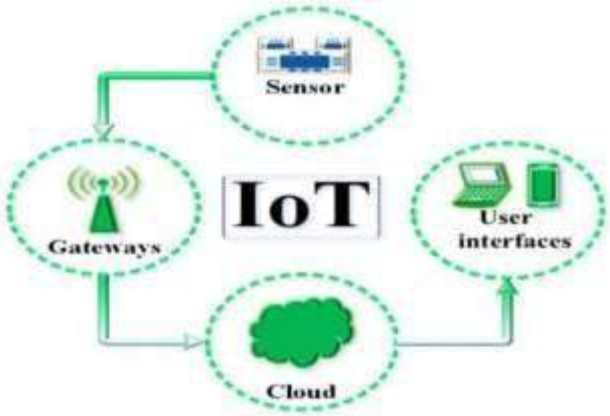
Modern technology offers much interactive design to ease complex tasks into simple touch panels controls. Multicolor touch panels have replaced hard switches in our household appliances and the trend is increasing for almost every smart home devices.

User interface design has higher significance in today’s competitive market, it often determines the user whether to choose a particular device or appliance. Users will be interested to buy new devices or smart gadgets if it is very user friendly and compatible with common wireless standards.

Components of IoT

The basic three components of IoT includes “Things”, “Internet”, and “Connectivity”.

IoT Components	Description	Example
Things	Physical objects	
	Sensors	It sense the physical environment
		e.g. Speed of a car
	Autuators	It affect the physical environment
		e.g. Brake controller of car
Internet	Platforms	Type of middleware used to connect IoT components (objects, people, services, etc) to IoT. Provide numerous functions:
		· Access to devices
		· Ensuring proper installation/behavior of device
		· Data analytics
		· Interoperable connection to local network, cloud or other devices.
	Services	e.g. Cloud services can be used to:
		· Process big data and turn it into valuable information
		· Build and run innovative applications
		· Optimize business processes by integrating device data
Connectivity	Network	IoT components are tied together by networks, using various wireless and wireline technologies, standards, and protocols to provide pervasive connectivity.
		· RFID



CHARACTERISTICS OF THE INTERNET OF THINGS :

There are the following characteristics of IoT as follows.

1. Connectivity is an important requirement of the IoT infrastructure. Things of IoT should be connected to the IoT infrastructure. Anyone, anywhere, anytime can connectivity should be guaranteed at all times Without connection, nothing makes sense.

2. Intelligence and Identity –

   The extraction of knowledge from the generated data is very important. For example, a sensor generates data, but that data will only be useful if it is interpreted properly. Each IoT device has a unique identity. This identification is helpful in tracking the equipment and at times for querying its status.

3. Enormous Scalability –

   The number of elements connected to the IoT zone is increasing day by day. Hence, an IoT setup should be capable of handling the massive expansion. The data generated as an outcome is enormous, and it should be handled appropriately.

4. Dynamic and Self-Adapting (Complexity) –

   IoT devices should dynamically adapt themselves to the changing contexts and scenarios. Assume a camera meant for the surveillance. It should be adaptable to work in different conditions and different light situations (morning, afternoon, night).

5. Architecture

   IoT architecture cannot be homogeneous in nature. It should be hybrid, supporting different manufacturers ' products to function in the IoT network. IoT is not owned by anyone engineering branch. IoT is a reality when multiple domains come together.

6. Safety

   There is a danger of the sensitive personal details of the users getting compromised when all his/her devices are connected to the internet. This can cause a loss to the user. Hence, data security is the major challenge. Besides, the equipment involved is huge. IoT networks may also be at the risk. Therefore, equipment safety is also critical.

Applications of IoT:
Before going to read about IoT applications , just watch this reference video-
https://youtu.be/91aXs9E0qAI

The concept of the Internet of Things entered our lives in 1999. However, in fact, the first IoT application has entered our lives before. In 1991, a system was designed to send images of the coffee machine to the computers of academics at the University of Cambridge three times a minute. Due to the fact that it is online and in real time, this system is considered to be the first application of the Internet of Things in the world.Some of the applications are-

1. Wearables
2. Connected cars/smart cars
3. Smart cities
4. Smart industries
5. Smart agriculture
6. Smart retail
7. Energy management
8. Smart healthcare
9. Smart poultry and farming
10. Smart dust

Smart City IoT Applications

Smart city IoT applications aim to ensure that citizens live in maximum comfort and resource consumption is made wisely. It aims to reduce and ultimately eliminate traffic density, air pollution, polluted water resources, garbage and waste problems, population agglomeration, and crime rates. In short, the goal of smart city IoT applications is basically to put an end to all problems that endanger human safety, health and well-being. Smart cities that solve the traffic problem with smart traffic lights or end the dirty water problem with clean water projects get very efficient results.

Smart Farming IoT Applications

Smart Farming IoT Applications

To understand **smart farming IoT applications**, first, let's define the concept of smart farming. Combining many advanced technologies and using them in agriculture is called smart farming. Smart farming and smart agriculture use modern informatics methods in agriculture and aim to increase productivity. Thanks to smart agriculture, the life of both producers and farmers is much easier.

Thanks to **IoT applications in agriculture**, control of agricultural areas can be done remotely. This saves time for everyone working in the agricultural sector. The simultaneous operation of agricultural machinery is one of the factors that save time and speed.

The use of **IoT based applications in agriculture** is also an action aimed at protecting the environment. With the spread of smart farming practices all over the world, it is aimed to prevent problems such as water scarcity and drought in time.

It is aimed to reduce the chemical products used in agriculture and thus to produce healthier products. Thus, the cost of such chemical products will be eliminated and savings will be provided.

IoT based applications in agriculture, which ensure that each natural resource is used only in the required amount, aim to avoid waste.

Smart Grids IoT Applications

It is aimed to establish mutual electronic communication between the supplier and the consumer through smart grids. Smart grids **IoT and its applications** work intertwined with each other. Smart grids IoT applications are encountered in many fields especially in the energy systems. It is aimed to add smart meters and monitoring systems to the electricity networks and thus to monitor a more reliable, quality and safe process.

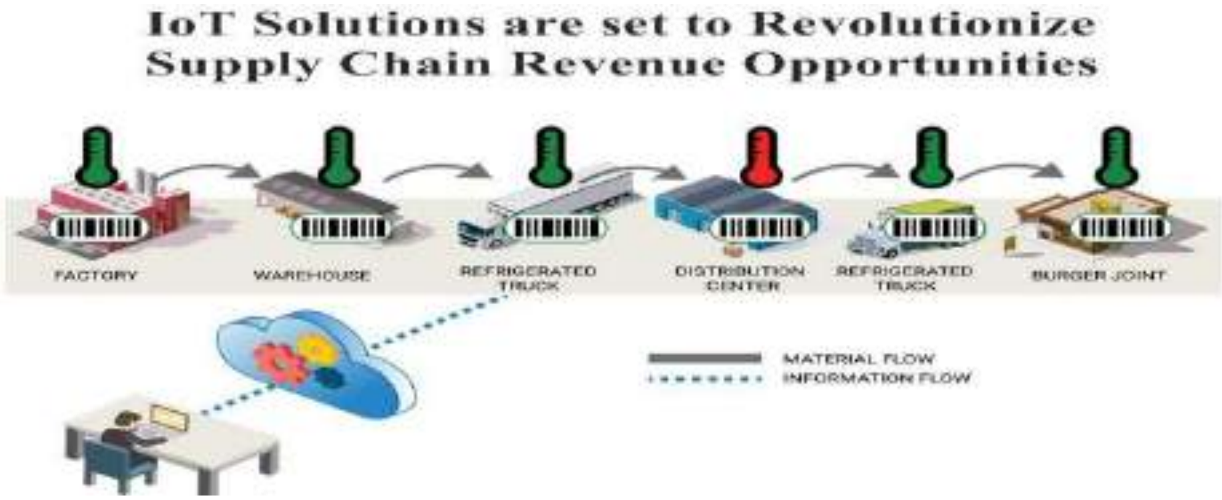
Smart Home IoT Applications

IoT applications used in smart homes and smart buildings are used to control the systems inside these buildings and homes. It provides control of systems that provide lighting, heating, security, alarm, entertainment systems and so on.

Smart Retail IoT Applications

Internet of Things technology develops IoT applications to improve in-store customer experience and provide a higher quality service. It brings customers, objects, sales processes and transactions to the digital platform.

Smart Supply Chain IoT Applications



One of the areas that IoT technology has entered into the digitalization process is the smart supply chain IoT applications. IoT technology is able to control the complexity caused by the increasing number of data and the increasing number of complex variables on a global scale.

Wearables IoT Applications

Wearable **IoT applications** are mainly used in the health and fitness sectors. Thanks to the wearable devices manufactured with IoT technology, it is possible to make measurements of people's body, disease follow-up and many other measurements at any time. The number of wearables IoT applications that are causing serious and positive changes especially in the health sector is increasing day by day.

Connected Car IoT Applications

Connected car **IoT based applications** used in transportation have resulted in many solutions such as smart traffic control, unmanned autonomous navigation, smart parking systems, and the establishment of digital communication between the vehicle and the driver.

Connected Health IoT Applications

Connected health **IoT applications ideas** in health and fitness have contributed greatly to the development of mobile digital medical systems. Many opportunities such as remote monitoring of patients 'health status, emergency notification systems, wearable IoT devices and monitoring of patients' body values have been realized thanks to the ideas of health and fitness IoT applications.

Industrial Internet IoT Applications

Industrial IoT applications, aka IIot Applications, are a set of applications that fundamentally restructure the industry. Therefore, these developments in technology are also called as a new industrial revolution. It is a system where all the smart devices used in production or other industrial fields can communicate with each other and control this communication from a single device.

Industrial IoT applications can be seen in many areas. Frequently used in industrial automation, smart robot systems, smart sensors, wearable technology integration, logistics, software, security, energy management.

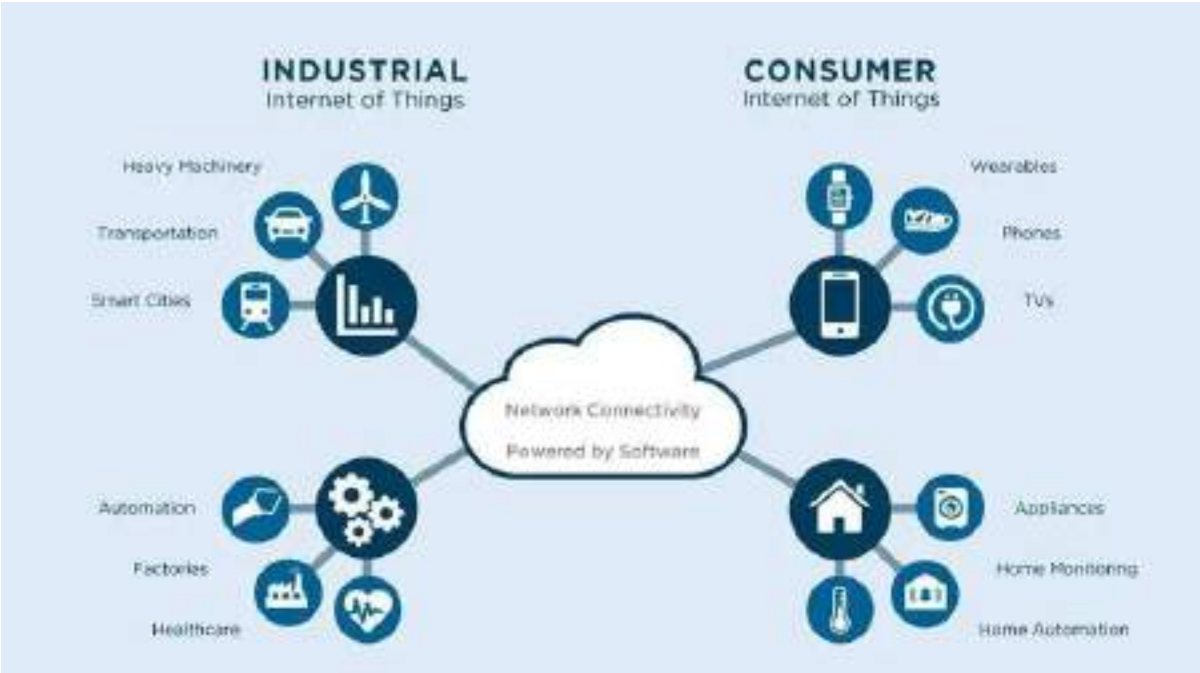
IOT CATEGORIES

Defining IoT with a consumer part and an industrial/business segment

The first distinction people started to make was between a consumer IoT and an Internet of Things for industrial applications or Industrial IoT as a way to distinguish between many types of IoT use cases and applications. Yet, as said and as with all terminology there were certainly overlaps in the definitions of these forms of IoT.

This is why some organizations and individuals, for instance, rather talk about the Internet of Everything, while others opt to drop the term IoT alltogether and mention it in terms of specific use cases and contexts such as smart cities, smart metering, smart buildings, smart office, smart wearables, Industrial Internet or smart homes, all of course with their own meaning and, again, with more subdivisions.

Consumer IoT (CIoT)



The Consumer Internet of Things or CIoT is where you will find applications and use cases to track your personal 'assets' *(asset tracking)*, from your pet to your skateboard. Or where you will find the connected 'smart appliances' such as connected refrigerators, washing machines, light bulbs, etc.

Also wearables for consumer use (wearables are also used in healthcare and in factories, to name just two) and all sorts of consumer electronics such as smart wristwear belong to this category, along with all sorts of smart home appliances like thermostats or connected parking door openers.

The applications get better and smarter. They also get more independent from other devices such as smartphones. This is certainly the case with smart wearables.

A simple definition of the Consumer Internet of Things is all we need: the Internet of Things as it's used for consumer applications and consumer-oriented services.

What is Consumer Internet of Things (CIoT)?

Consumer IoT (CIoT) refers to the use of IoT for consumer applications and devices. Common CIoT products include smartphones, wearables, smart assistants, home appliances, etc.

Typically, CIoT solutions leverage Wi-Fi, Bluetooth, and ZigBee to facilitate connectivity. These technologies offer short-range communication suitable for deployments in smaller venues, such as homes and offices.

Typically, in Consumer IoT, data volumes and data communication needs are low and limited. That's why there are many technologies of which some are specifically designed for consumer applications, ranging from smart home connectivity standards to special operating systems for wearables.

IIoT (Industry 4.0)

The Industrial Internet of Things or IIoT describes typical industry use cases across a range of sectors. Two examples of Industrial IoT use cases: predictive maintenance and asset management. Some people see the Industrial Internet of Things more in a context of 'heavy' industries like manufacturing or utilities. But it is also used for use cases in, for example smart cities.

If we look at it as a sort of 'Business Internet of Things' it is clear that there are some overlaps with the Consumer Internet of Things. For instance: if you have a smart thermostat and smart energy consumption meter in your house they are on one hand consumer applications because they are for personal usage.

But from the perspective of the company that uses it to send you invoices and to help optimize energy consumption it is a business matter (e.g., 'smart grid'). So, the terms are not that good but that's how it is and it's better to look at use cases than at these broad categories because just as there are many different applications in the Consumer Internet of Things, there are also many in IIoT and some are hard to compare. Most industrial IoT applications relate to the digital transformation of manufacturing or to the rise of smart industry though.

IoT Enablers and connectivity layers

IoT Enablers:

- **RFIDs:** uses radio waves in order to electronically track the tags attached to each physical object.
- **Sensors:** devices that are able to detect changes in an environment (ex: motion detectors).
- **Nanotechnology:** as the name suggests, these are extremely small devices with dimensions usually less than a hundred nanometers.
- **Smart networks:** (ex: mesh topology).

The most basic architecture is a three-layer

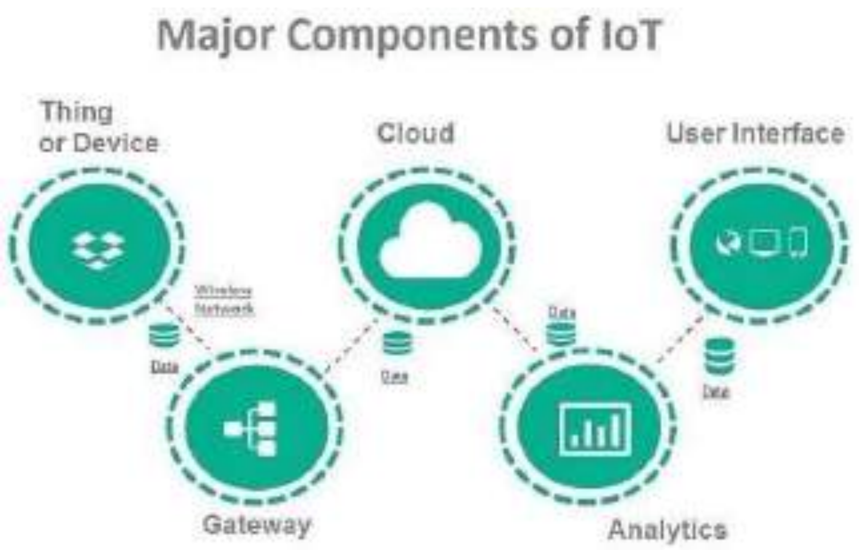
It was introduced in the early stages of research in this area. It has three layers, namely, the perception, network, and application layers.

(i) The perception layer is the physical layer, which has sensors for sensing and gathering information about the environment. It senses some physical parameters or identifies other smart objects in the environment.

(ii) The network layer is responsible for connecting to other smart things, network devices, and servers. Its features are also used for transmitting and processing sensor data.

(iii) The application layer is responsible for delivering application specific services to the user. It defines various applications in which the Internet of Things can be deployed, for example, smart homes, smart cities, and smart health.

What are the major components of Internet of Things



How the devices are connected?

Most of the modern smart devices and sensors can be connected to low power wireless networks like Wi-Fi, ZigBee, Bluetooth, Z-wave, LoRAWAN etc… Each of these wireless technologies has its own pros and cons in terms of power, data transfer rate and overall efficiency.

Developments in the low power, low cost wireless transmitting devices are promising in the area of IoT due to its long battery life and efficiency. Latest protocols like 6LoWPAN- IPv6 over Low Power Wireless Personal Area Networks have been adapted by many companies to implement energy efficient data transmission for IoT networks.

6LoWPAN uses reduced transmission time (typically short time pulses) and thus saves energy

.

Lecture Plan

Subjet- Energy Conversion -|| Semester- 5th		
1	Topic	Alternators , Types of alternators & their constructional features
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Define Alternator 2. Explain the construction of an alternator 3. Classify alternators
3	Tools used	Chalk, Blackboard and Projector
4	Lecture Description	An alternator is an electrical generator that converts mechanical energy to electrical energy in the form of alternating current. ... In principle, any AC electrical generator can be called an alternator, but usually the term refers to small rotating machines driven by automotive and other internal combustion engines. • Construction of Alternator The construction of an alternator consists of field poles placed on the rotating fixture of the machine. An alternator is made up of two main parts: a rotor and a stator. The rotor rotates in the stator, and the field poles get projected onto the rotor body of the alternator. • Stator The stator is the stationary part of a rotary system, found in electric generators, electric motors, sirens, mud motors or biological rotors In an electric motor, the stator provides a magnetic field that drives the rotating armature; in a generator, the stator converts the rotating magnetic field to electric current. The alternator stator attaches to the outside of the crankcase such that an assembly of iron poles, each with a coil of wire wound around it, fits inside the rotor. As the crankshaft spins, magnets are constantly reversing the magnetism in each pole. •Rotor The Rotor is also known as the rotating magnetic core, or Flywheel, of an Alternator. Rotors are the moving part in an Alternator that have permanent magnets that move around the Stator's iron plates to generate an Alternating Current (AC). Opening the alternator reveals a large cylinder with triangular finger poles around the circumference. This is the rotor. A basic alternator is made up of a series of alternating finger pole pieces placed around coil wires called field windings that wrap around an iron core on the rotor shaft. •Armature Winding Armature winding in an alternator may be either closed type open type. Closed winding forms star connection in armature winding of alternator. There are some common properties of armature winding. ... Winding is so arranged in different armature slots, that it must produce sinusoidal emf. •Types of Alternators

		There are 1- Pulsed radar 2- Continuous wave radar The pulse radar transmits short rectangular pulses & the continuous wave radar transmits continuous sinusoidal EM waves. Pulse Radar The Radar, which operates with pulse signal is called the Pulse Radar. Pulse Radars can be classified into the following two types based on the type of the target it detects. • Basic Pulse Radar • Moving Target Indication Radar Continuous Wave Radar The Radar, which operates with continuous signal or wave is called Continuous Wave Radar. They use Doppler Effect for detecting non-stationary targets. Continuous Wave Radars can be classified into the following two types. • Unmodulated Continuous Wave Radar • Frequency Modulated Continuous Wave Rada
5	Questions	1. What do you mean by alternator ? 2. Explain the construction of alternator with diagram.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	Alternators , Basic working principle of Alternator & the relation between speed & frequency
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1.Explain principle of working of Alternator 2.State the relation between speed & frequency

3	Tools used	Chalk, Blackboard and Projector

4 5	Lecture Description	See below

•Working principle of Alternator

An alternator operates on the same fundamental principle of electromagnetic induction as a DC generator. The working of an alternator is based on the principle that when the flux linking a conductor changes, an emf is induced in the conductor.

Like a DC generator, an alternator also has an armature winding and a field winding. But there is one important difference between the two.

In a DC generator, the armature winding is placed on the rotor in order to provide a way of converting alternating voltage generated in the winding to a direct voltage at the terminals through the use of a rotating commutator.

•Speed & Frequency

When we discuss the term frequency, we are referring to how many cycles can occur in one second. Frequency is measured in hertz (shout out to Heinrich Hertz) or CPS (cycles per second). Two factors affect the frequency in an alternator: rotation speed and the number of poles.

As the armature rotates through the field, it starts to create a waveform (as we saw in the last chapter). One full mechanical rotation of the armature creates one full sine wave on a two-pole alternator. If the two-pole alternator spins three complete revolutions in one second, it will create three full sine waves in that one second. We would say that the frequency is at three cycles per second or three hertz.

A machine's rotational speed is measured in rotations per minute or RPM. However, we are not concerned with minutes, but rather, with seconds when dealing with frequency. Therefore, RPM must be converted to rotations per second (RPS). As there are 60 seconds in a minute, all we have to do is to divide the RPM by 60 to convert it to RPS.

For example, if the armature is spinning at a rate of 1800 RPM on a two-pole alternator, we can say that it is spinning at 30 rotations per second. If this alternator has two poles, then in one second it will generate 30 cycles of voltage. It then could be said to have a frequency of 30 cycles per second or 30 Hertz. The frequency of an alternator is directly proportional to the rotational speed of the alternator.

•Number of poles

If we add poles to the alternator, we can change the frequency. In a two-pole alternator, one side of the armature passes from north to south, and then south to north, to create one complete sine wave. I f we add two more poles, as in Figure 54, then Side A of the armature will move past two north poles and two south poles in one full mechanical revolution.

Two full sine waves are created in one complete mechanical revolution. If a two-pole alternator creates one cycle of voltage in one second (or one hertz of frequency), a four pole alternator will create two cycles of voltage in one second (or two hertz).

The frequency of an alternator is directly proportional to the number of poles in the alternator. Knowing that rotation speed is directly proportional to frequency and that the number of poles is directly proportional to frequency, we can use a formula. The formula looks like this:

$$f=PN/120$$

Where ,P=total no. of magnetic poles

N=speed of rotor in rpm or synchronous speed

f =frequency of generated e.m.f. in Hz

	Problems-	
5	Questions	1. Explain working principle of an alternator with diagram. 2. State the relationship between speed and frequency.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-3

Subject - Energy Conversion -|| Semester- 5th		
1	Topic	Alternators , Terminology in armature winding & Expressions for winding factors (Pitch factor , Distribution factor)
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Define turn , coil & winding 2.Define pitch factor & derive expression for pitch factor 3.Define distribution factor & derive expression for distribution factor
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	**•Terms used in Alternators** Some of the basic terms related to the Armature Winding are defined as follows: Turn: A turn consists of two conductors connected to one end by an end connector. Coil: A coil is formed by connecting several turns in the series. Winding: A winding is formed by connecting several coils in series. **•Pitch factor and Expression for pitch factor** **In short pitched coil, the induced emf of two coil sides get vectorially added and give resultant emf of the loop. In short pitched coil, the phase angle between the induced emf of two opposite coil sides is less than 180o (electrical). But we know that, in full pitched coil, the phase angle between the induced emf of two coil sides is exactly 180° (electrical).** **Hence, the resultant emf of a full pitched coil is just the arithmetic sum of the emfs induced on both sides of the loop. We well know that vector sum or phasor sum of two quantities is always less than their arithmetic sum. The pitch factor is the measure of resultant emf of a short-pitched coil in comparison with resultant emf of a full pitched coil.** Hence, it must be the ratio of phasor sum of induced emfs per coil to the arithmetic sum of induced emfs per coil. Therefore, it must be less than unity.

Let us assume that, a coil is short pitched by an angle α (electrical degree). Emf induced per coil side is E. The arithmetic sum of induced emfs is 2E. That means, 2E, is the induced voltage across the coil terminals, if the coil would have been full pitched.

Now, come to the short pitched coil. Resultant emf of the short pitched coil

$$\text{Resultant emf} = 2E\cos\alpha/2$$

Now, as per definition of pitched factor,

$$K_p = \frac{\text{Voltage sum of the induced emfs per coil}}{\text{Arithmetic sum of the induced emfs per coil}}$$

$$= \frac{2E\cos\alpha/2}{2E} = \cos\alpha/2$$

•Distribution factor & Expression for distribution factor

If all the coil sides of any one phase under one pole are bunched in one slot, the winding obtained is known as concentrated winding and the total emf induced is equal to the arithmetic sum of the emfs induced in all the coils of one phase under one pole.

But in practical cases, for obtaining smooth sinusoidal voltage waveform, armature winding of alternator is not concentrated but distributed among the different slots to~~ form polar g~~roups under each pole. In distributed winding, coil sides per phase are displaced from each other by an angle equal to the angular displacement of the adjacent slots. Hence, the induced emf per coil side is not an angle equal to the angular displacement of the slots.

So, the resultant emf of the winding is the phasor sum of the induced emf per coil side. As it is phasor sum, must be less than the arithmetic sum of these induced emfs.

Resultant emf would be an arithmetic sum if the winding would have been a concentrated one.

As per definition, distribution factor is a measure of resultant emf of a distributed winding in compared to a concentrated winding.

We express it as the ratio of the phasor sum of the emfs induced in all the coils distributed in some slots under one pole to the arithmetic sum of the emfs induced. Distribution factor is,

		Emf with distributed winding Kd = Emf with concentrated winding
5	Questions	1. Define winding. 2. Differerentiate between pitch factor and distribution factor.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-4

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	Alternators , Explain harmonics , it's causes and impact on winding factor
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Define harmonics 2. Explain causes & impacts of harmonics on winding factor
3	Tools used	Chalk, Blackboard and Projector
4	Lecture Description	**•Harmonics** **We know that in synchronous machines or alternators, the voltage and currents are induced. These voltage and currents are sinusoidal waveforms. ....These sinusoidal components or harmonic functions are called harmonics of the complex wave.** **•Causes of harmonics on winding factor** This pitch factor is the fundamental component of emf. The flux wave may consist of space field harmonics also, which give rise to the corresponding time harmonics in the generated voltage waveform. A 3rd harmonic component of the flux wave, may be imagined as produced by three poles as compared to one pole for the fundamental component.

		In the view of this, the chording angle for the rth harmonic becomes r times the chording angle for the fundamental component and pitch factor for the rth harmonic is given as, $$Kpr = \cos \alpha r/2$$ The rth harmonic becomes zero, if, $$\cos \alpha r/2 = 0 \text{ or } \alpha r/2 = 90°$$ In 3 phase alternator, the 3rd harmonic is suppressed by star or delta connection as in the case of 3 phase transformer. Total attention is given for designing a 3 phase alternator winding design, for 5th and 7th harmonics. For 5th harmonic $5\ \alpha/2 = 90°$ or, $= 180°/5 = 36°$ For 7th harmonic $7\ \alpha/2 = 90°$ or,. $\alpha = 180°/7 = 25.7°$ Hence, by adopting a suitable chording angle of $\alpha = 30°$, we make most optimized design armature winding of alternator.
5	Questions	1. What is harmonics ? What is the effect of harmonics on winding factor. 2. How harmonic affect the winding factor.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-5

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	Alternators , E.M.F. equation of an alternator
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Derive the e.m.f. equation of an alternator
3	Tools used	Chalk, Blackboard and Projector
4	Lecture Description	**•Equation of induced e.m.f.**

		The emf induced by the alternator or synchronous generator is three-phase alternating in nature. Let us derive the mathematical equation of emf induced in the alternator. Let, Z = number of conductors in series per phase. Z = 2T, where T is the number of coils or turns per phase. One turn has two coil sides or conductor . P = Number of pole f = frequency of induced emf in Hertz Φ = flux per pole in webers. Kp = pitch factor, Kd = distribution factor, Kf = Form factor N = Speed of the rotor in rpm(revolutions per minute) N/60 = Speed of the rotor in revolutions per second. Time taken by the rotor to complete one revolution, dt = 1/(N/60)= 60/N second In one revolution of the rotor, the total flux Φ cut the by each conductor in the stator poles, dΦ = ΦP weber By faraday's law of electromagnetic induction, the emf induced is proportional to rate of change of flux. Average induced emf Equation Of Alternator $d\Phi/dt = \Phi P/60/N$. $= \Phi NP/60$ We know, the frequency of induced emf $f = PN/120$ or $N = 120f/P$ Submitting the value of N in the induced emf equation, We get $\Phi P/60 \times N = \Phi P/60 \times 120f/P$ $= 2\Phi f$ If there are Z conductors in series per phase,

		Average induced emf Equation Of Alternator $= 2\Phi f Z = 4\Phi f T$ RMS value of emf per phase = Form factor x Average value of induced emf = 1.11 x 4 Φ f Tp RMS value of emf per phase = 4.44 Φ f T volts The obtained above equation is the actual value of the induced emf for full pitched coil or concentrated coil. However, the voltage equation gets modified because of the winding factors. Actual induced emf per phase = 4.44 Kp Kd Φ f T volts = 4 Kf Kp Kd Φ f T volts
5	Questions	Derive an expression for induced emf in an alternator.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-6

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	Alternators , Explain armature reaction & its effect on emf at different power factor of load
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Define armature reaction 2. Explain effect of armature reaction on emf at different power factor of load
3	Tools used	Chalk, Blackboard and Projector
4	Lecture Description	**•Armature Reaction** Armature reaction is an important aspect in DC generator and AC synchronous Generator or Alternator. Armature reaction is defined as the effect of armature flux on the main flux produced by the field poles. When the load is applied to the alternator, current starts flowing through the armature conductors. Since the current is alternating in nature, it induces a flux in the conductor, called as armature flux.

The armature flux thus produced will react with the main field flux and distort the effect of main flux. Due to this distortion, the main flux will either strengthen or weaken.
The distortion may depend on the type of load applied to the alternator. There are three types of distorting effects.
Cross magnetizing effect
Demagnetizing effect
Magnetizing effect

Armature reaction at unity power factor load

When a resistive load with unity power factor is connected to the alternator, the load current will starts to flow through the armature conductors. As the load is resistive, the armature current will be in phase with the induced voltage.

The armature current will produce its own flux in the conductor, which will also be in phase with the induced voltage. Since the induced emf lags behind the main flux by 900, the armature flux produced will also be delayed by 900 with respect to the main flux.

Armature reaction at zero power factor lagging load

When an inductive load with zero lagging power factor is connected to alternator, the load current starts to flow through the armature conductors.

The load current will be delayed by 900 and so the armature flux produced will also be shifted by 900 with respect to the poles.

There will be a phase difference of 900 between the armature flux and main field flux. It can be seen that, the armature flux will be in direct opposition to the main flux.

Waveform and phasor diagram for demagnetizing effect

Thus the main flux gets decreased in this loading condition. This effect of armature reaction on this load is said to be demagnetizing effect.

Due to this, the main flux gets weaken and so the emf induced will be reduced. To maintain the same value of generated emf, filed excitation will have to be increased to overcome the demagnetizing effect.

Armature reaction at zero power factor leading load

		When a capacitive load with zero leading power factor is connected, the load current starts to flow through the armature conductors. In this load condition, the load current will be advanced by 900 and so the flux produced will also be advanced by 900 with respect to emf induced. So the armature flux will be in phase with the main field flux, resulting in strengthening of the field flux. Thus the main flux gets increased in this loading condition. Waveform and phasor diagram for magnetizing effect The armature reaction in this load is said to be magnetizing effect. Due to this effect, the main flux gets weaken and so the emf induced will be reduced. To maintain the same value of generated emf, filed excitation will have to be reduced to overcome the magnetizing effect. For any intermediate power factor, the effect of armature reaction in alternator will be partly distorting and partly demagnetizing.
5	Questions	1. Define armature reaction. 2. Explain effects of armature reaction on different power factor of load. 3. Draw the vector diagrams for unity power factor,zero p.f. lagging,& zero p.f. leading.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-7

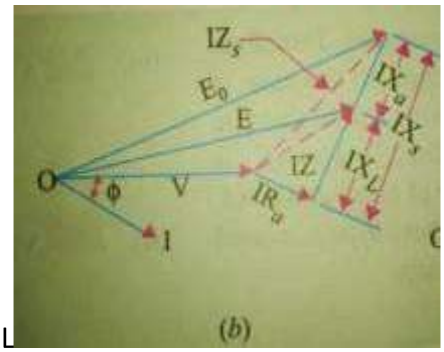
Subject -Energy Conversion -|| Semester- 5th		
1	Topic	Alternators , The vector diagram of loaded alternator
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1.Draw vector diagrams of loaded alternator
3	Tools used	Chalk, Blackboard and Projector

4	Lecture Description	**•Synchronous reactance and Vector diagrams of loaded alternator**

•Synchronous reactance and Vector diagrams of loaded alternator

For the same field excitation, terminal voltage is decreased from its no-load value E0 to V (for a lagging power factor). This is because of

1. drop due to armature resistance, IRa



2. drop due to leakage reactance, IXL

3. drop due to armature reaction.

The drop in voltage due to armature reaction may be accounted for by assumiung the presence of a fictitious reactance Xa in the armature winding. The value of Xa is such that IXa represents the voltage drop due to armature reaction. The leakage reactance XL (or XP) and the armature reactance Xa may be combined to give synchronous reactance XS.

Hence XS =XL Xa

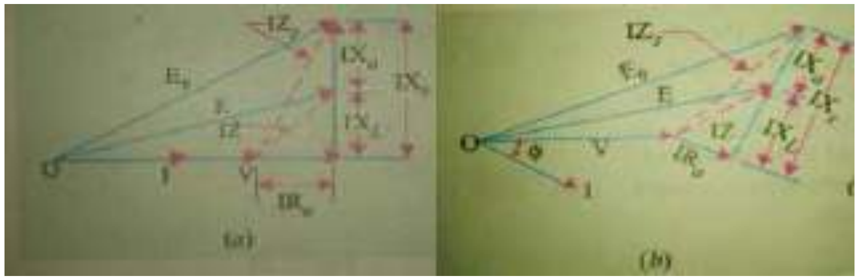
Therefore, total voltage drop in an alternator under load is = IRa jIXS = I(Ra jXS) = IZS where ZS is known as synchronous impedance of the armature, the word 'synchronous' being used merely as an indication that it refers to the working conditions.

Hence, we learn that the vector difference between no-load voltage E0 and terminal voltage V is equal to IZS, as shown in Fig. 37.26.

•Vector Diagrams of a Loaded Alternator

Before discussing the diagrams, following symbols should be clearly kept in mind.

E0 = No-load e.m.f. represents the maximum value of the induced e.m.f.

		E = Load induced e.m.f. It is the induced e.m.f. after allowing for armature reaction. E is vectorially less than E0 by IXa. Sometimes, it is written as Ea.  V=Terminal voltage , It is vectorially less than Eo by IZs or it is vectorially less than E by Iz , where , Z=√(,it may also be written as Za. I=Armature current/phase and In Fig.(a) is shown the case for unity p.f., in Fig. (b) for lagging p.f. and in Fig. 37.27 (c) for leading p.f. All these diagrams apply to one phase of a 3-phase machine. Diagrams for the other phases can also be drawn similarly.
5	Questions	Define synchronous reactance.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-09

Subject -Energy Conversion -|| Semester- 5^{th}		
1	Topic	Alternators , Testing of alternators 1.Open circuit test 2.Short circuit test
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Explain Why alternator testing was done ? 2. Describe the open circuit test of alternators . 3. Describe the short circuit test of alternators.
3	Tools used	Chalk, Blackboard and Projector

4	Lecture Description	**•Open circuit test of alternators** The open-circuit test or the no-load test, is performed by driving the generator at its rated speed while the armature winding is left open. The field current is varied in suitable steps and the corresponding values of the open-circuit voltage varied in suitable steps and corresponding values of the open-circuit voltage between any two pair of terminals of the armature windings are recorded. The OCC follows a straight-line relation as long as the magnetic circuit of the synchronous generator does not saturate. In the linear region, most of the applied mmf is consumed by the air-gap; the straight line is appropriately called the air-gap line. As the saturation sets in, the OCC starts deviating from the air-gap line. **•Short circuit test of alternators** The short-circuit test provides information about the current capabilities of a synchronous generator. It is performed by driving the generator at its rated speed when

		the terminals of the armature winding are shorted. An ammeter is placed in series with one of the three shorted lines. The field current is gradually increased and the corresponding value of the armature current is recorded. When the per phase short-circuit current is plotted as a function of the field current, the graph is called the short circuit characteristic of a generator.
5	Questions	1. Explain need of testing of alternators ? 2.How open circuit test of alternators was done ? 3.How short circuit test of alternators was done ?
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

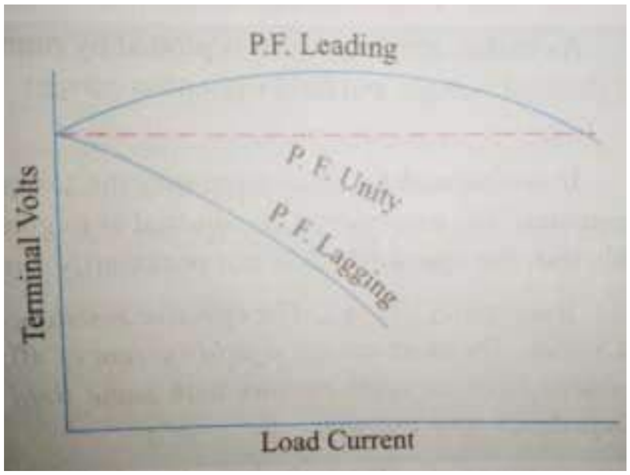
Lecture Plan-10

Subject -Energy Conversion -|| Semester- 6th		
1	Topic	Alternators , Determination of voltage regulation of alternator by direct loading & synchronous impedance method
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Define voltage regulation. 2. Explain determination of voltage regulation , step by step.
3	Tools used	Chalk, Blackboard and Projector

4 Lecture Description

•Voltage regulation

the voltage regulation of an alternator is defined as the change in its terminal voltage when full load is removed, keeping field excitation and speed constant, divided by the rated terminal voltage. The value of the regulation not only depends on the load current but also on the power factor of the load.



% regulation 'up' = Eo—V/V ×100

•Determination of voltage regulation

In the case of small machines, the regulation may be found by direct loading.

The alternator is driven at synchronous speed and the terminal voltage is adjusted to its rated value V.

The load is varied until the wattmeter and ammeter (connected for the purpose) indicate the rated values at desired p.f. Then the entire load is thrown off while the speed and field excitation are kept constant.

The open-circuit or no-load voltage E0 is read.

% regn = Eo—V/V ×100

V = Rated terminal voltage

E0= No load induced e.m.f.

2. In the case of large machines, the cost of finding the regulation by direct loading becomes prohibitive.

•Direct loading method

The direct loading method is well suited for small rating alternators without actual loading. The below shows the connection diagram of the alternator to which three-phase load is connected across the armature terminals.

Voltage Regulation of Alternator

The field excitation is adjusted until the voltmeter across the armature terminal reads the rated voltage. Now the load is connected and increased until the ammeter reads the rated current i.e., full-load armature current. When the load is connected the terminal voltage decreases due to armature reaction effect. Again by increasing the field excitation, the voltage V is adjusted to its rated value.

Now keeping the same speed and field excitation the entire load is disconnected from the armature terminals. Thus the load current becomes zero and the voltmeter reads the value of induced emf Eo i.e., no-load terminal voltage without armature reaction drop. Once Eo is obtained regulation is determined.

•Synchronous impedance method

The synchronous impedance method or EMF method is one of the methods of finding voltage regulation of alternator or synchronous generator. In order to calculate regulation by this method, it requires the following data,

The per-phase armature resistance Ra,

The open-circuit test data shows the relationship of the characteristics between open-circuit voltage and field current, and

The short-circuit characteristic is the graph of short-circuit current against field current.

Open-Circuit Test :

The circuit diagram for the open-circuit test is shown below. Let the alternator is running at synchronous speed driven by the prime mover with no-load connected across the armature. The voltmeters connected across the armature open-circuited lines measure open-circuit voltage.

The rheostat in the field circuit (rotor) is set to its maximum value.

Now increase the field current by varying the field rheostat so that flux produced by field winding increases.

The increased flux increases induced emf in the alternator thereby increasing the terminal voltage of the alternator.

The various values of open-circuit line voltage Voc at different field current If values are noted

The below curve shows the open-circuit test characteristics drawn between field current and open-circuit voltage.

Short circuit test:

A short-circuit test is performed by running the alternator at rated speed. The below shows the circuit diagram with short-circuited armature terminals. The ammeters connected in the armature circuit measures short-circuit current.

Synchronous Impedance or EMF Method

Before short-circuiting the armature terminals, the field current is reduced to a minimum value by increasing the resistance of the field rheostat.

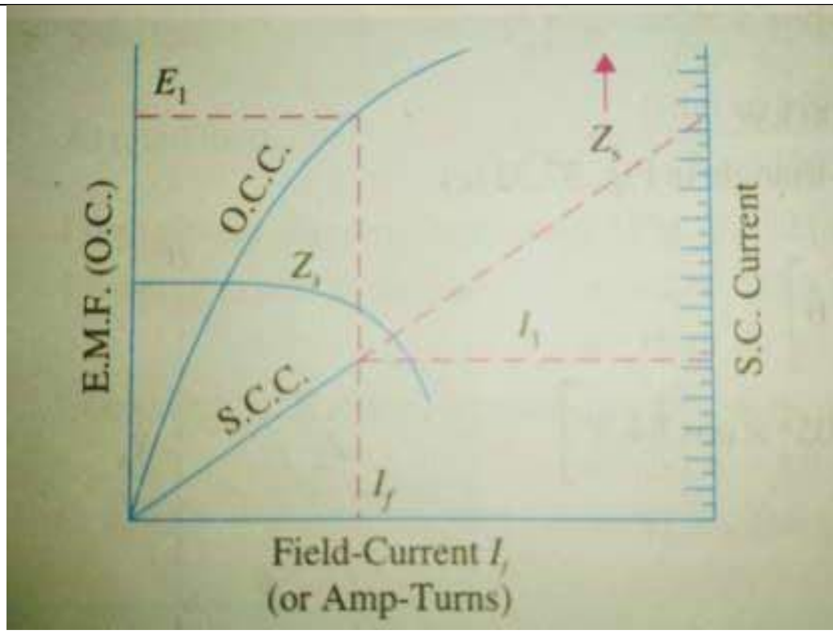
Now, the field current is gradually increased till the armature winding reads full load current.

The full load current is obtained through armature ammeter reading.

The various short-circuit current Isc values at a various field current Ia values are noted.

The relation between both values gives short-circuit characteristics of an alternator as shown below.

Calculation of Synchronous Impedance from O.C.C. and S.C.C. :



At a particular value of field current If from open-circuit and short-circuit test the ratio of open-circuit voltage to the short-circuit current gives the synchronous impedance Zs. Now at a field current If, the armature open-circuit voltage is Eoc when the winding is short-circuited, the terminal voltage is zero. Hence, the short-circuit current Isc against synchronous impedance Zs is circulated in the armature circuit by using whole the voltage Eoc. That is,.

Eoc = Isc Zs

Zs=OC voltage /SC current

Zs =Eoc/Isc

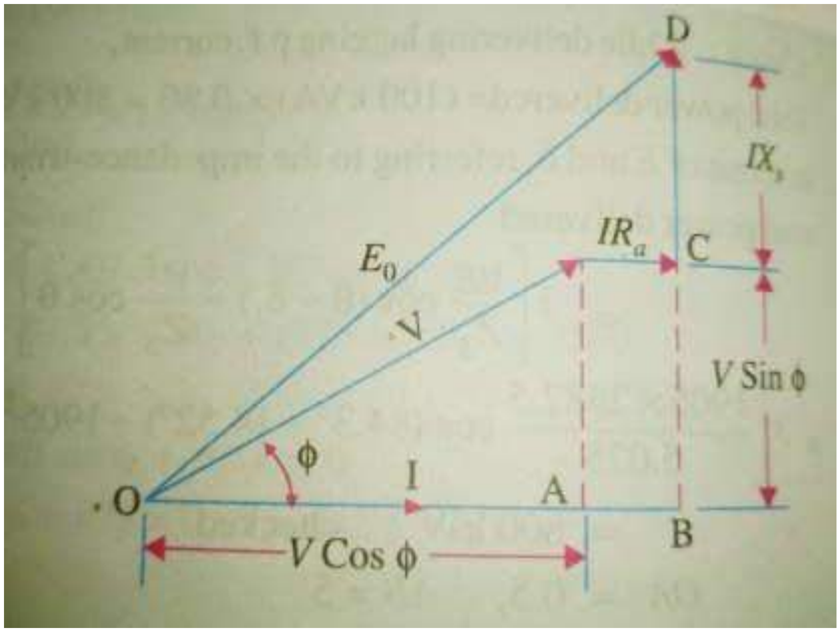
From the above, the synchronous reactance is given by,

$$X_s = \sqrt{Z_s \times Z_s - R_s \times R_s}$$

The above figure shows the O.C.C. and S.C.C. characteristics.

Referring to the above figure, at a field current value say OA, the alternator terminal voltage or induced emf will be AB. From the short-circuit curve with the same field current the short-circuit current in the armature windings will be AC. Therefore, from the above O.C.C. and S.C.C. curves, the synchronous impedance Zs will be, Zs = AB volts/ AC amps

Regulation Calculation :



By determining the values of per-phase armature resistance Ra and synchronous reactance Xs, the vector diagram can be drawn for any load and any power factor as shown below.

From the vector diagram with a lagging power factor.

$$(OD)^2 = (OB)^2 + (BD)^2$$

$$Eo^2 = (V\cos + IRa)^2 + (V\sin + IXs)^2$$

In general, Eo is given by,

$$Eo = \sqrt{(V\cos\phi + IRa)^2 + (V\sin\phi + IXs)^2}$$

Where,

Vph = Rated phase voltage

Ia = Phase value of current depending upon the load condition.

cos φ = Power factor of the load

Positive (+) sign for lagging power factor

Negative (-) sign for leading power factor

Therefore the formula for % regulation is given as,

		$\%regn = E_o - V / V. \times 100$ It will be found that all these methods differ chiefly in the way the no-load voltage E0 is found in each case.

5	Questions	1. Explain process of determining of voltage regulation . 2. Differentiate direct loading method and synchronous impedance method .
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-11

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	Alternators , Parallel operation of alternators using syncroscope & dark and bright lamp method
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Say why parallel operation of alternators was done ? 2. Explain the methods of parallel operation of alternators.
3	Tools used	Chalk, Blackboard and PPT

4 Lecture Description

Syncroscope method

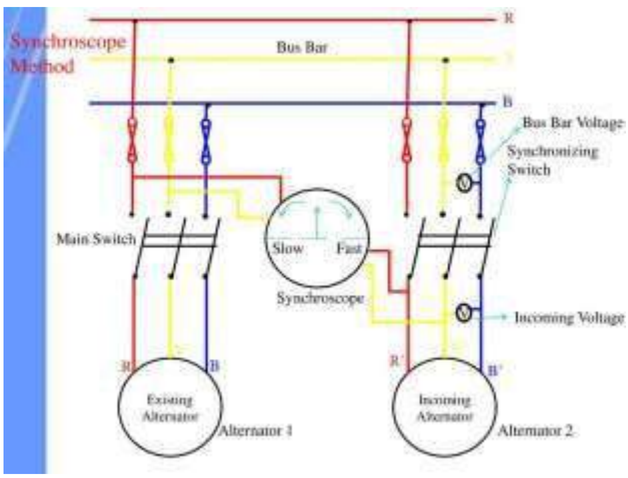
A synchroscope is a device with a pointer rotating on a dial. The pointer indicates the phase relation between two alternators by its direction of rotation and also tells whether the incoming alternator is running faster or slower than the existing alternator.

The following must be the essential conditions for synchronization :

The generated terminal voltage of the incoming alternator must be equal to the existing alternator or bus-bar voltage.

The frequency at which the voltage is generated must be also equal

There must be zero phase difference with equal phase sequences



At first, the switch is kept open and if the frequencies of two alternators are different then there will be a movement in the pointer placed on the moving coil. The rotational movement is in such a way that,

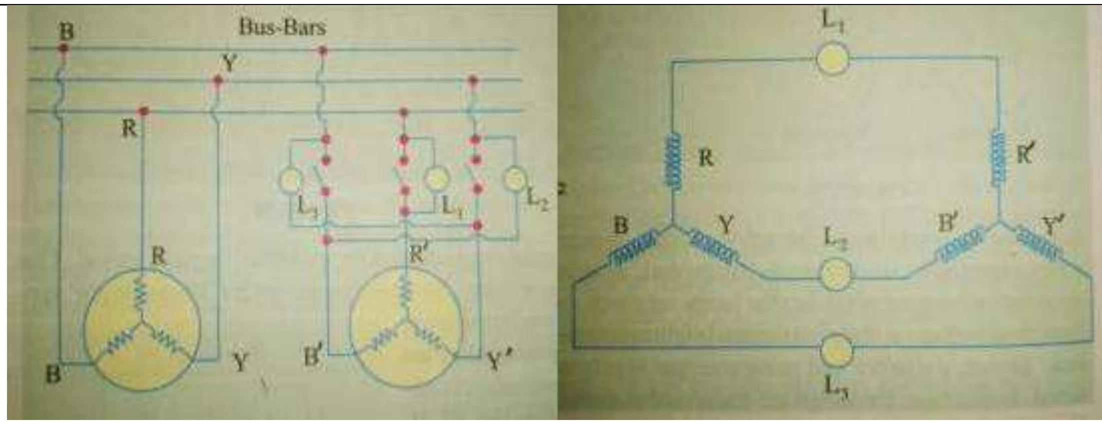
If the frequency of the incoming alternator is low compared to the bus-bars or existing alternator then the rotation is in the anti-clockwise direction.

If the frequency of the incoming alternator is more, then the rotation is in the clockwise direction.

If the difference in frequencies is more the speed of rotation is also more. By adjusting the driving speed of the incoming alternator the frequency can be adjusted i.e., by adjusting the speed of the prime mover. The pointer remains stationary in a vertical position when both the frequencies are made equal. Thus the switch S is closed at this instant and the incoming alternator is made to run parallelly with other alternators.

•Two Bright and One Dark Lamp Method

The connections for this method are shown in figure below and it is useful in finding whether the alternator frequency is lower or higher than the bus bar frequency.

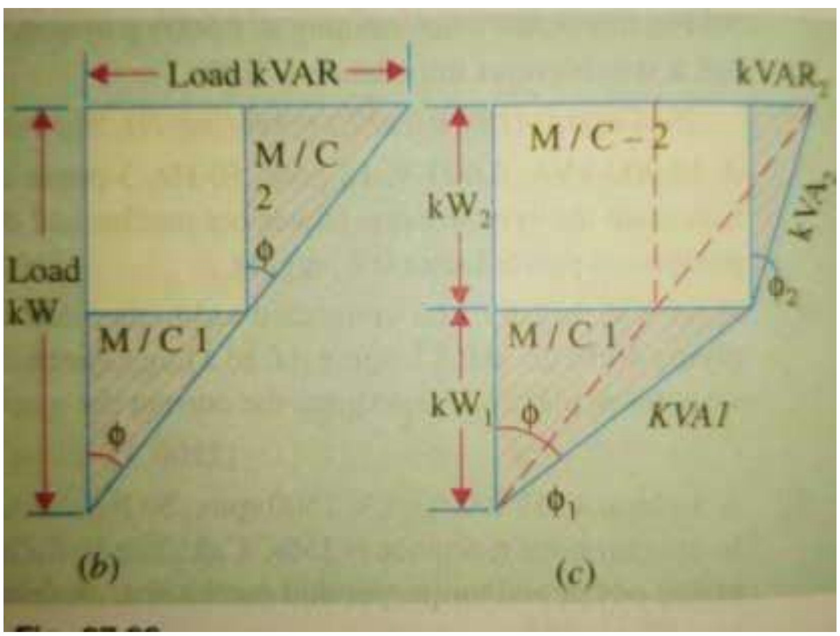
		 Here, the lamp L2 is connected across the pole in the middle line of synchronizing switch as similar to the dark lamp method, whereas the lamps L1 and L3 are connected in a transposed manner. The voltage condition checking is similar to the previous method and after it, the lamps glow bright and dark one after another. The lower or higher value of alternator frequency in comparison with bus bar frequency is determined by the sequence in which the lamps become dark and bright. The sequence of becoming bright and dark L1- L2 – L3 indicates that the incoming generator frequency is higher than the bus bar frequency. Hence, the alternator speed has to be reduced by prime mover control till the flickering rate is brought down to a small. On the other hand, the sequence flickering L1- L3 – L2 indicates that incoming alternator frequency is less than that of bus bar. Hence, the speed of the alternator is increased by the prime mover till the rate of flickering is brought down to as small as possible. The synchronizing switch is then closed at the instant when lamps L1 and L3 are equally bright and lamp L2 is dark. The disadvantage of this method is that the correctness of phase sequence cannot be checked. However, this requirement is unnecessary for permanently connected alternators where checking of phase sequence is enough to be carried out for the first time of operation alone.
5	Questions	1. Define synchronisation. 2. State the conditions for parallel operation of alternators. 3. Explain two bright and one dark lamp method with diagram. 4. Explain synchroscope method with diagram.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-12

<table>
<tr><td colspan="3">Subject. -Energy Conversion -||
Semester- 5th</td></tr>
<tr><td>1</td><td>Topic</td><td>Alternators , Distribution of load by parallel connected alternators.</td></tr>
<tr><td>2</td><td>Intended Learning Outcome</td><td>On successful completion of the Lecture students are able to:
Explain distribution of load by parallel connected alternators.</td></tr>
<tr><td>3</td><td>Tools used</td><td>Chalk, Blackboard & Projector</td></tr>
<tr><td>4</td><td>Lecture Description</td><td>

•Distribution of load by parallel connected alternators

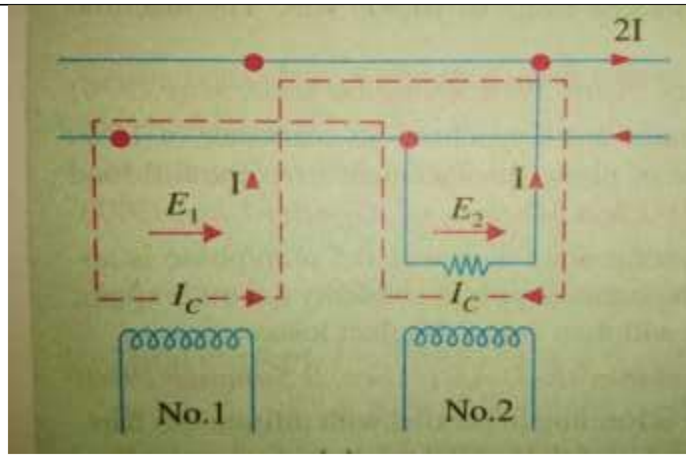
When several alternators are required to run in parallel, it probably happens that their rated outputs differ. In such cases it is usual to divide the total load between them in such a way that each alternator takes the load in the same proportion of its rated load in total rated outputs. The total load is not divided equally. Alternatively, it may be desired to run one large alternator permanently on full load, the fluctuations in load being borne by one or more of the others. If the alternators are sharing the load equally the power triangles are as shown in Fig



Sharing of load when two alternators are in parallel

Consider two alternators with identical speed load characteristics connected in parallel as shown in Fig:

</td></tr>
</table>

		 Let E1, E2 be the induced emf per phase, Z1, Z2 be the impedances per phase, I1, I2 be the current supplied by each machine per phase Z be the load impedance per phase, V be the terminal voltage per phase From the circuit we have $V = E1 - I1Z1 = E2 - I2Z2$ and hence, $I1 = E1 - V/Z1$ and $I2 = E2 - V/Z2$ and also $V = (I1 + I2)\, Z = IZ$ solving above equations $I1 = [(E1- E2)\, Z + E1\, Z2]/ [\, Z(Z1 + Z2) + Z1Z2]$ $I2 = [(E2- E1)\, Z + E2\, Z1]/ [\, Z(Z1 + Z2) + Z1Z2]$ The total current $I = I1 + I2 = [E1Z2 + E2Z1] / [\, Z(Z1 + Z2) + Z1Z2]$ And the circulating current or synchronizing current $Is = (E1 - E2) / (Z1 + Z2)$
5	Questions	Explain distribution of load by parallel connected alternators with diagram.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-13

Subject -Energy Conversion -|| Semester- 5th	
1 Topic	Synchronous motors , Constructional features of synchronous motors

<table>
<tr><td>2</td><td>Intended Learning Outcome</td><td>On successful completion of the Lecture students are able to:
1. Explain details of construction of synchronous motors.</td></tr>
<tr><td>3</td><td>Tools used</td><td>Chalk, Blackboard and Projector</td></tr>
<tr><td>4</td><td>Lecture Description</td><td>

Construction of synchronous motors

A synchronous motor is generally made up of two parts, a stator the stationary part of the machine that carries the armature winding in which the voltage is generated, and a rotor the rotating part of the machine that produces the main field flux. Of course, this motor has other parts and components. But electric motors are generally known by these two parts. In the following, we will get acquainted with all these components and parts, and we will mention the function and role of each in the motor.

1. Stator Construction

The stationary part of the machine is called Stator. It includes various parts like stator frame, stator core, stator windings, and cooling arrangement. They are explained below in detail.

1.1. Stator Frame

It is the outer body of the machine made of cast iron, and it protects the inner parts of the machine.

1.2. Stator Core

The stator core is made of silicon steel material. It is made from many stamps which are insulated from each other. Its function is to provide an easy path for the magnetic lines of force and accommodate the stator winding.

1.3. Stator Winding

Slots are cut on the inner periphery of the stator core in which 3 phase or 1 phase winding is placed. Enameled copper is used as a winding material. The winding is star connected. The winding of each phase is distributed over several slots. When the current flows in a distributed winding it produces an essentially sinusoidal space distribution of EMF.

2. Rotor Construction

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</table>

The rotating part of the machine is called the Rotor. There are two types of rotor construction, namely the salient pole type and the cylindrical rotor type. In the following, we will deal with each of these two types of rotors and you can find their differences in their structure and operation.

2.1. Salient Pole Rotor

The term salient means projecting. Thus, a salient pole rotor consists of poles projecting out from the surface of the rotor core. The end view of a typical 6 pole salient pole rotor is shown below in the figure.

Since the rotor is subjected to changing magnetic fields, it is made of steel laminations to reduce eddy current losses. Poles of identical dimensions are assembled by stacking laminations to the required length. A salient pole synchronous machine has a non-uniform air gap. The air gap is minimized under the pole centers and it is maximum in between the poles.

They are constructed for medium and low speeds as they have a large number of poles. A salient pole generator has a large diameter. The salient pole rotor has the following important parts.

Non-Salient Pole Rotor or Cylindrical Rotor

In this type of rotor, unlike Salient Pole Rotor, there are no projected poles, but the poles are formed by the current flowing through the rotor exciting winding. Cylindrical rotors are usually made from solid forgings of high-grade nickel chrome-molybdenum steel. It has a comparatively small diameter and long axial length.

This type of Rotor is useful in high-speed machines. The cylindrical rotor has two or four poles on the rotor. Such a construction provides greater mechanical strength and allows more accurate dynamic balancing. The smooth rotor of the machine makes fewer windage losses and the operation is less noisy because of the uniform air gap. Cylindrical Rotors are driven by steam or gas turbines. The machines are built in many ratings from 10 MVA to over 1500 MVA. The biggest size used in India has a rating of 500 MVA installed in the super thermal power plant.

Cylindrical Rotors have the following seven parts and components. We have listed all these parts and components below and have prepared a brief and useful explanation for each of them for more familiarity.

1) Rotor Core

The rotor core is made of silicon steel stampings. The rotor core is placed on the shaft (we will explain what the shaft is). At the outer periphery, slots are cut in which exciting coils are placed.

2) Rotor Winding (Exciting Winding)

It is placed on the rotor slots, and the current is passed through the winding in such a way that the poles are formed according to the requirement.

3) Pole Core and Pole Shoe

It is made of laminated sheet steel material. The Pole core provides the least reluctance path for the magnetic field and the pole shoe distributes the field over the whole periphery uniformly to produce a sinusoidal wave.

4) Spider

It is made of cast iron to provide an easy path for the magnetic flux. It is keyed to the shaft and at the outer surface, pole core and pole shoe are keyed to it.

5) Field Winding (Exciting Winding)

It is wound on the former and then placed around the pole core. DC supply is given to it through slip rings. When direct current flow through the field winding, it produces the required magnetic field.

6) Damper Winding

At the outermost periphery, holes are provided in which copper bars are inserted and short-circuited at both sides by rings forming Damper winding.

7) Rings

Slip rings provide DC supply to the rotor windings.

3. Miscellaneous Parts and Components

We mentioned at the beginning of this section that in addition to the rotor and stator and their components, there are other components in a synchronous motor. In the following, you will become more familiar with these parts and their role. The miscellaneous parts are given below.

A) Brushes

		Brushes are made of carbon, and they slip over the slip rings. A DC supply is given to the brushes. Current flows from the brushes to the slip rings and then to the exciting windings. **B) Bearings** **Bearings are provided between the shaft and the outer stationary body to reduce friction. They are made of high carbon steel.** **C) Shaft** **The shaft is made of mild steel. Mechanical power is taken or given to the machine through the shaft.**
5	Questions	1. Explain construction of synchronous motors.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-14

UNIT DISCUSSION- I

1. Why rotor slots of an induction motor are skewed ?
2. What is plugging ?
3. Write down the advantages of stationary armature of alternator ?
4. What do you mean by voltage regulation of an alternator ?
5. What is damper bar or damper winding ?
6. State the applications of synchronous motor.
7. How the direction of rotation of a single phase induction motor can be reversed ?
8. Why parallel operation of alternators needed ?
9. What is hunting ?
10. What is V-curves of synchronous motor ?
11. What do you mean by infinite bus ?

12. Which type of alternator is used in hydroelectric power plants and why ?

13. What is the relation between electrical degrees and mechanical angle of an alternator ?

14. State the difference between coil pitch and pole pitch.

15. State pitch factor and distribution factor.

16. What is the function of damper bars ?

17. What is the principle of operation of synchronous motor ?

18. Why a starting winding is needed ?

19. In which rotor high starting torque is produced and why ?

20. What are the types of alternator ?

21. What frequency voltage must be applied to a 20-pole synchronous motor so that it runs at a speed of 300 rpm ?

22. Can a three phase IM never runs at synchronous speed ?

23. How frequency of rotor current is related to slip ?

24. Why are alternator rated in kVA , not in kW ?

25. Why single phase induction motor is not self starting ?

26. If a single phase induction motor fails to start but gives a humming sound , what is the reason behind it ?

27. What are the two advantages of short pitched coil in an alternator wdg.

28. Calculate the value of 'Kd' for a 36-slots , 4-poles single layer three phase alternator.

29. For traction purpose which motors are used ?

30. State two applications of DC series motor.

31. State the difference between coil pitch and pole pitch.

32. State pitch factor and distribution factor.

33. What is the function of damper bars ?

34. What is the principle of operation of synchronous motor ?

35. Why a starting winding is needed ?

36. In which rotor high starting torque is produced and why ?

37. What are the types of alternator ?

38. What frequency voltage must be applied

Lecture Plan-15

<table>
<tr><td colspan="3">Subject -Energy Conversion -||
Semester- 5th</td></tr>
<tr><td>1</td><td>Topic</td><td>Synchronous motors , Principles of operation & concept of load angle</td></tr>
<tr><td>2</td><td>Intended Learning Outcome</td><td>On successful completion of the Lecture students are able to:
1. Describe principle of operation of synchronous motor.
2. Brief idea about concept of load angle.</td></tr>
<tr><td>3</td><td>Tools used</td><td>Chalk, Blackboard and PPT</td></tr>
<tr><td>4</td><td>Lecture Description</td><td><h2>Principle of operation</h2>
Synchronous motors are a doubly excited machine, i.e., two electrical inputs are provided to it. Its stator winding which consists of a We provide three-phase supply to three-phase stator winding, and DC to the rotor winding.
The 3 phase stator winding carrying 3 phase currents produces 3 phase rotating magnetic flux. The rotor carrying DC supply also produces a constant flux. Considering the 50 Hz power frequency, from the above relation we can see that the 3 phase rotating flux rotates about 3000 revolutions in 1 min or 50 revolutions in 1 sec.
At a particular instant rotor and stator poles might be of the same polarity (N-N or S-S) causing a repulsive force on the rotor and the very next instant it will be N-S causing attractive force.
But due to the inertia of the rotor, it is unable to rotate in any direction due to that attractive or repulsive forces, and the rotor remains in standstill condition. Hence a synchronous motor is not self-starting.
Here we use some mechanical means which initially rotates the rotor in the same direction as the magnetic field to speed very close to synchronous speed. On achieving synchronous speed, magnetic locking occurs, and the synchronous motor continues to rotate even after the removal of external mechanical means. </td></tr>
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<tr><td></td><td></td><td>
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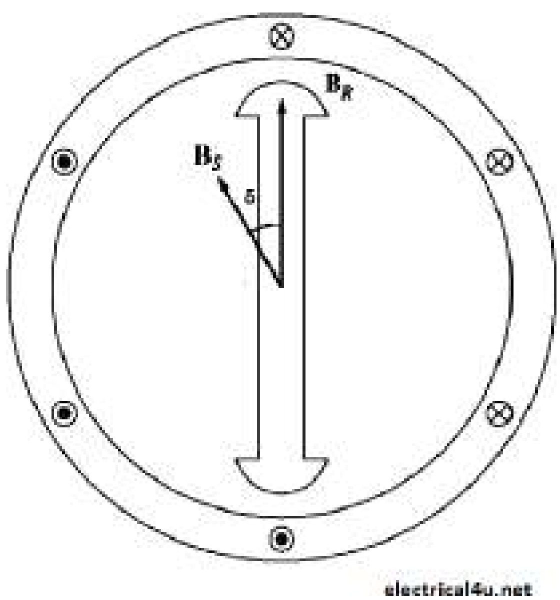
Here we use some mechanical means which initially rotates the rotor in the same direction as the magnetic field to speed very close to synchronous speed. On achieving synchronous speed, magnetic locking occurs, and the synchronous motor continues to rotate even after the removal of external mechanical means.

Concept of load angle

Load angle is nothing but an angle different between stator axis and rotor pole axis of the synchronous motor. For ideal motor, the load angle is zero since the rotor poles aligned with stator poles, but in practice, this is not possible. The motor has both mechanical and electrical losses, hence load angle is always present in the synchronous motor.



Load angle
</td></tr>
<tr><td>5</td><td>Questions</td><td>1. Explain principle of operation of synchronous motor.
2. What do you mean by load angle ?</td></tr>
<tr><td>6</td><td>Reference</td><td>https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.htmlhttps://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html</td></tr>
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Lecture Plan-16

<table>
<tr><td colspan="3">Subject -Energy Conversion -||
Semester- 5th</td></tr>
<tr><td>1</td><td>Topic</td><td>Synchronous motors , Derive power torque ,power developed</td></tr>
<tr><td>2</td><td>Intended Learning Outcome</td><td>On successful completion of the Lecture students are able to:
1. Explain different power torques of synchronous motors.
2. Describe expression for power developed.</td></tr>
<tr><td>3</td><td>Tools used</td><td>Chalk, Blackboard and Projector</td></tr>
<tr><td>4</td><td>Lecture Description</td><td>

•Starting Torque:
The torque is being developed at the starting time of the motor. It is also called as breakaway torque. The starting torque of the synchronous motor is purely depending on the method of starting the motor.

•Running Torque:
The full load torque of the motor is called running torque. The running torque is defending on the motor specifications.

•Pull-in Torque:
Let we assume the synchronous motor is started and the speed is nearer to the synchronous speed, during that time the stator pulls the rotor into synchronism, that torque is called pull-in torque.

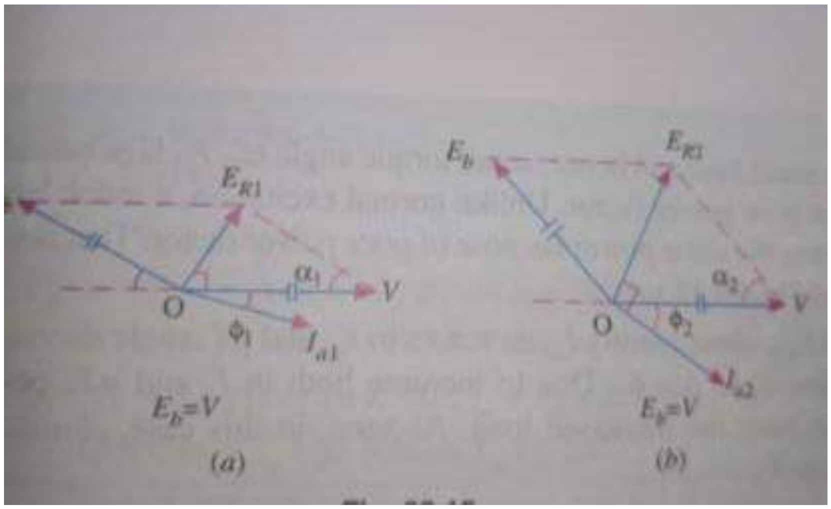
•Pull out Torque:
Let we assume the motor is running at the maximum torque, beyond that slight increase in load causes the motor pulls out the synchronism, that maximum torque is called pull out torque. The pull out torque will be three to four-time of the full load torque of the motor.

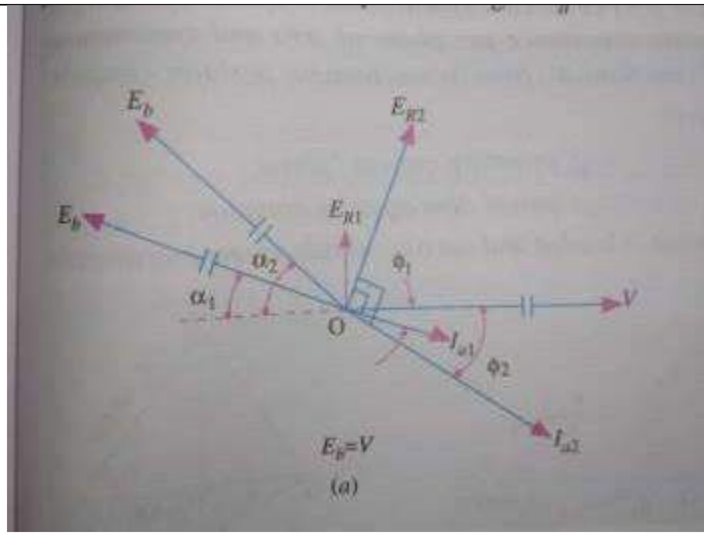
•Expression for Power developed
Except for very small machines , the armature resistance of a synchronous motor is negligible as compared to its synchronous reactance. Hence , the equivalent circuit for the motor becomes as shown in figure. From the phasor diagram of figure (b) , it is seen that
$AB = E_b \sin \alpha = I_a X_s \cos\varphi$
or ,. $V I_a \cos\varphi = E_b V / X_s \sin \alpha$
Now , $V I_a \cos\varphi$ = Motor power input/phase
:. $P_{in} = E_b V / X_s \sin \alpha$
$= 3 E_b V / X_s \sin \alpha$
Since stator Cu losses have been neglected , Pin also represents the gross mechanical power (Pm) developed by the motor.
:. $P_m = 3 E_b V / X_s \sin \alpha$
</td></tr>
</table>

		The gross torque developed by the motor is $T_g = 9.55 P_m / N_s$. N-m........... N_s in rpm .l
5	Questions	1. Define power torque.Explain different types of torque. 2. Derive expression for power developed by a synchronous motor.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-17

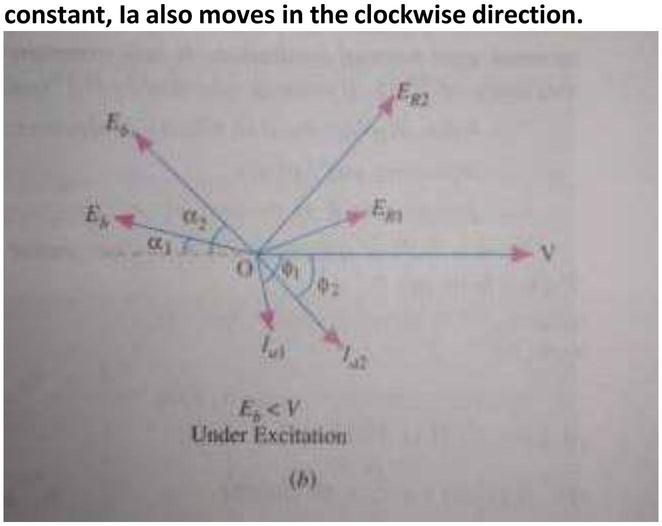
Subject -Energy Conversion -|| Semester- 5th		
1	Topic	Synchronous motors , Effect of varying load with consant excitation.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Calculate the velocity of Geostationary Satellite. 2. Calculate the round trip time delay of geostationary satellite.
3	Tools used	Chalk, Blackboard & Projector

4	Lecture Description	**With normal excitation, $E_b = V$, with under excitation, $E_b < V$ and with over-excitation, $E_b > V$. Whatever the value of excitation, it would be kept constant during our discussion. It would also be assumed that R_a is negligible as compared to X_S so that phase angle between E_R and I_a i.e., $\theta = 900$.** **(i) Normal Excitation**  **Fig. (a) shows the condition when motor is running with light load so that (i) torque angle α_1 is small (ii) so E_{R1} is small (iii) hence I_{a1} is small and (iv) φ_1 is small so that cos φ_1 is large.** **Now, suppose that load on the motor is increased as shown in Fig. (b). For meeting this extra load, motor must develop more torque by drawing more armature current. Unlike a d.c. motor, a synchronous motor cannot increase its I_a by decreasing its speed and hence E_b because both are constant in its case. What actually happens is as under :** **1. Rotor falls back in phase i.e., load angle increases to α_2 as shown in Fig. (b),** **2. The resultant voltage in armature is increased considerably to new value E_{R2},** **3. as a result, I_{a1} increases to I_{a2}, thereby increasing the torque developed by the motor,** **4. φ_1 increases to φ_2, so that power factor decreases from cos φ_1 to the new value cos φ_2. Since increase in I_a is much greater than the slight decrease in power factor, the torque developed by the motor is increased (on the whole) to a new value sufficient to meet the extra load put on the motor. It will be seen that essentially it is by increasing its I_a that the motor is able to carry the extra load put on it. A phase summary of the effect of increased load on a synchronous motor at normal excitation is shown in Fig. (a) It is seen that there is a comparatively much greater increase in I_a than in φ.**



(ii) Under excitation

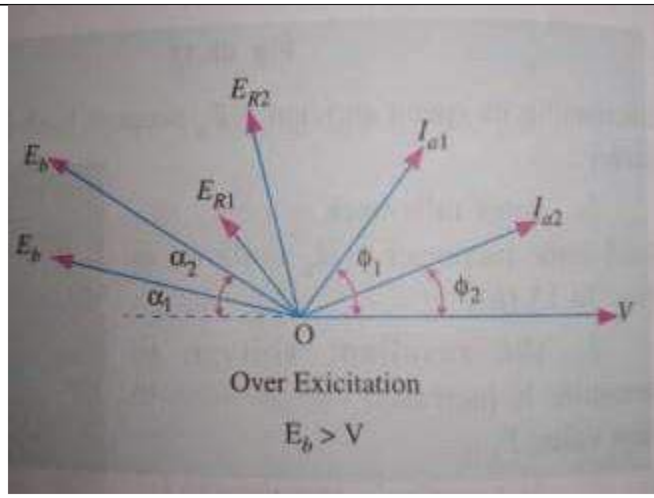
When the excitation is reduced in such a way that Eb < V, the motor is said to 'Under Excited'. The resultant emf Er advanced in a clockwise direction and as an angle θ is constant, Ia also moves in the clockwise direction.



As seen from figure (b) angle φ is increased and the power factor is decreased. To maintain Ia cos φ constant, Ia is increased. Therefore for low values of excitation, Ia increases and the power factor is lagging in nature.

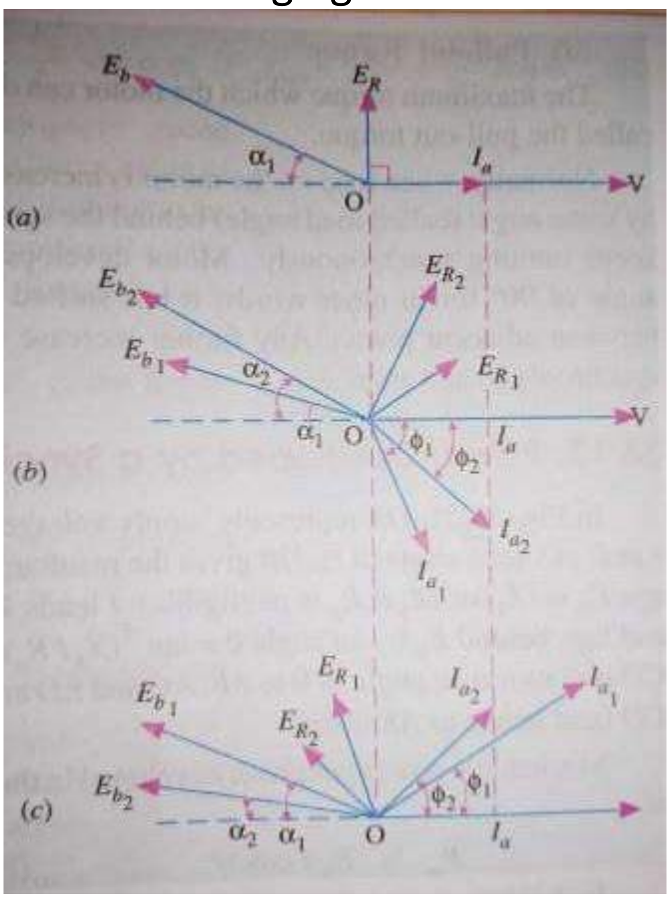
(iii) Over Excitation

When excitation is increased in such a way that Eb > V, the motor is said to be 'Over-Excited'. The resultant emf Er is moved in the anti-clockwise direction and so the Ia also moves in an anti-clockwise direction (as an angle between Er and Ia is θ). Current Ia leads voltage by an angle φ as shown in figure.

		 As excitation goes on increasing, Er and Ia go on moving in the anti-clockwise direction, angle φ increases. power factor decreases and Ia increases as Ia cos φ = constant. Thus for high values of excitation, current Ia increases and power factor is leading in nature.
5	Questions	1. Write notes on effect of increased load with normal , under & over excitation.

6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

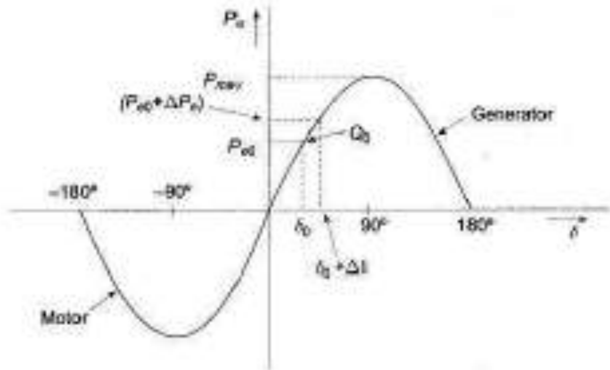
Lecture Plan-18

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	Synchronous motors , Effect of varying excitation with constant load.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Explain effect of changing excitation on constant load.
3	Tools used	Chalk, Blackboard and Projector
4	Lecture Description	**Effect of changing excitation on constant load**  As shown in Fig. (a), suppose a synchronous motor is operating with normal excitation (Eb = V) at unity p.f. with a given load. If Ra is negligible as compared to XS, then Ia lags ER by 90º and is in phase with V because p.f. is unity. The armature is drawing a power of V.Ia per phase which is enough to meet the mechanical load on the motor. Now, let us discuss the effect of decreasing or increasing the field excitation when the load applied to the motor remains constant. **(a) Excitation Decreased**

		As shown in Fig. (b), suppose due to decrease in excitation, back e.m.f. is reduced to Eb1 at the same load angle α1. The resultant voltage ER1 causes a lagging armature current Ia1 to flow. Even though Ia1 is larger than Ia in magnitude it is incapable of producing necessary power VIa for carrying the constant load because Ia1 cos φ1 component is less than Ia so that VIa1 cos φ1 < VIa. Hence, it becomes necessary for load angle to increase from α1 to α2. It increases back e.m.f. from Eb1 to Eb2 which, in turn, increases resultant voltage from ER1 to ER2. Consequently, armature current increases to Ia2 whose in-phase component produces enough power (VIa2 cos φ2) to meet the constant load on the motor. **(b) Excitation Increased** The effect of increasing field excitation is shown in Fig. (c) where increased Eb1 is shown at the original load angle α1. The resultant voltage ER1 causes a leading current Ia1 whose in-phase component is larger than Ia. Hence, armature develops more power than the load on the motor. Accordingly, load angle decreases from α1 to α2 which decreases resultant voltage from ER1 to ER2. Consequently, armature current decreases from Ia1 to Ia2 whose in-phase component Ia2 cos φ2 = Ia. In that case, armature develops power sufficient to carry the constant load on the motor. Hence, we find that variations in the excitation of a synchronous motor running with a given load produce variations in its load angle only.As shown in Fig. (a), suppose a synchronous motor is operating with normal excitation (Eb = V) at unity p.f. with a given load. If Ra is negligible as compared to XS, then Ia lags ER by 90º and is in phase with V because p.f. is unity. The armature is drawing a power of V.Ia per phase which is enough to meet the mechanical load on the motor. Now, let us discuss the effect of decreasing or increasing the field excitation when the load applied to the motor remains constant.
5	Questions	1. Explain effect of increased & decreased excitation on constant load.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

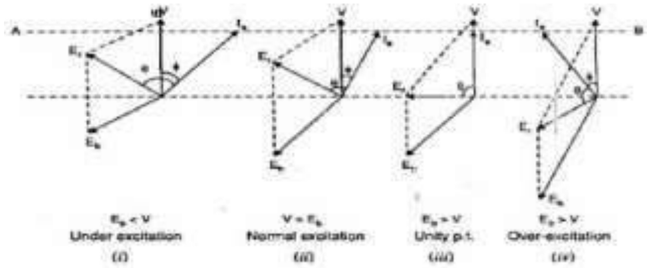
Lecture Plan-19

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	Synchronous motors , Power angle characteristics of cylindrical rotor motor
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Describe Power angle characteristics of cylindrical rotor motor.
3	Tools used	Chalk, Blackboard and Projector
4	Lecture Description	**Power angle characteristics**

		 Fig. 12.6 Power angle curve In a synchronous machine, if power angle is positive, it means the rotor is leading the rotating magnetic field and it is acting as a generator. The rotor is being driven by a mechanical torque. The power delivered is proportional to the sine of this angle. Maximum power is delivered at 90 degrees. If the power angle is negative, rotor is lagging the rotating magnetic field. it is acting as a motor, i.e. consuming power and producing mechanical torque. At -90 degrees, maximum torque is developed.
5	Questions	1. Write power angle characteristics of cylindrical rotor motor.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-21

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	Synchronous motors , Effect of excitation on armature current & power factor
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Define excitation. 2. Describe effect of excitation on armature current & power factor.
3	Tools used	Chalk, Blackboard and Projector

4	Lecture Description	 **Effect of excitation on armature current** the normal excitation, the armature current again increase and the operating power factor increases. However, the power factor is leading under this condition. Hence, the synchronous motor draws a leading current, when it is over excited and is equivalent to a capacitive load. **Effect of excitation on power factor** If the synchronous motor is under-excited, it has a lagging power factor. As the excitation is increased, the power factor improves till it becomes unity at normal excitation. Under such conditions, the current drawn from the supply is minimum.
5	Questions	1. What is the effect of excitation on armature current. 2. What is the effect of excitation on power factor.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-22

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	Synchronous motors , Hunting in synchronous motors.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Define hunting.
3	Tools used	Chalk, Blackboard and Projector
4	Lecture Description	

What Is Hunting In Synchronous Motor

We come across the term HUNTING when we study about three phase synchronous motor operations. The word hunting is used because after the sudden application of load the rotor has to search or 'hunt' for its new equilibrium position. That phenomenon is referred to as hunting in a synchronous motor. Now let us know what is the condition of equilibrium in synchronous motor.

A steady state operation of synchronous motor is a condition of equilibrium in which the electromagnetic torque is equal and opposite to load torque. In steady state, rotor runs at synchronous speed thereby maintaining a constant value of torque angle (δ). If there is a sudden change in load torque, the equilibrium is disturbed and there is resulting torque which changes the speed of the motor.

Unloaded synchronous machine has zero degree load angle. On increasing the shaft load gradually load angle will increase. Let us consider that load P1 is applied suddenly to unloaded machine shaft so machine will slow down momentarily.

Also load angle (δ) increases from zero degree and becomes $\delta1$. During the first swing electrical power developed is equal to mechanical load P1. Equilibrium is not established so rotor swings further. Load angle exceeds $\delta1$ and becomes $\delta2$. Now electrical power generated is greater than the previous one. Rotor attains synchronous speed. But it does not stay in synchronous speed and it will continue to increase beyond synchronous speed. As a result of rotor acceleration above synchronous speed the load angle decreases. So once again no equilibrium is attained. Thus rotor swings or oscillates about new equilibrium position. This phenomenon is known as hunting or phase swinging. Hunting occurs not only in synchronous motors but also in synchronous generators upon abrupt change in load.

Causes of Hunting in Synchronous Motor

		Sudden change in load. Sudden change in field current. A load containing harmonic torque. Fault in supply system. **Effects of Hunting in Synchronous Motor** It may lead to loss of synchronism. Produces mechanical stresses in the rotor shaft. Increases machine losses and cause temperature rise. Cause greater surges in current and power flow. It increases possibility of resonance
5	Questions	1. What do you mean by hunting in synchronous motors ? 2. What are the causes and effects of hunting in synchronous motors ?
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-23

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	Synchronous motors , Function of damper bars in synchronous motors & generators
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Define damper bars. 2. What are the functions of damper bars ?
3	Tools used	Chalk, Blackboard and Projector
4	Lecture Description	**Damper Windings are made using copper bars which are short-circuited at both ends. They act as a short circuit coils. Damper windings are placed in the face of the rotor poles of the synchronous machine along the pole axis.** **Function of damper bars in synchronous generators** The damper winding is used in the synchronous alternator to suppress or eliminate hunting. When in an alternator load is sudden changes then the rotor hunts and it tries to go a new equilibrium position. In other words, when the load in an alternator sudden changes(suppose it increases) then the rotor falls out of the synchronism. During this time rotor rotates less than synchronous speed. If we use a damper winding(basically a short circuit coil), an emf will be induced in the damper winding when the rotor rotates at less than synchronous speed.

Therefore the damper winding will create a flux which will increase the speed of the rotor to catch the synchronous speed. So the damper winding always tries to keep the rotor speed at the synchronous speed even during sudden load changes in the alternator.

Thus we can eliminate hunting in the alternator using damper winding. Damper winding helps to the alternator for smoothly running.

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Function of damper bars in synchronous motors

The purpose of using damper winding in the alternator and synchronous motor is not the same. The damper winding is used in the synchronous motor to provide starting torque.

We know that the synchronous motor is not self-starting. So to start the synchronous motor we follow various methods of starting. The method of starting using the damper winding is one of them.

Suppose a synchronous motor has damper winding. Initially, when the motor is off, the rotator is constant. When we give the main power supply to the stator winding of the synchronous motor, it creates a rotating magnetic field. Now the stator flux will cut the damper winding conductor and an emf will be induced in the damper winding as the induction motor principle.

So, initially, the synchronous motor will start as an induction motor and it will continue running after giving the auxiliary DC supply to the main field winding. When the rotor catches the synchronous speed there will be no relative speed between the stator and rotor and the damper winding will stop inducing emf. So it is clear that the damper winding works only starting time of the motor.

Thus we can start the synchronous motor using damper winding.

These were the reasons Why Damper Winding is used in Synchronous Machine.

5	Questions	1. What do you mean by damper bars or damper winding ? 2. What is the function of damper bars in synchronous motors ? 3. What is the function of damper bars in synchronous generators ?
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-24

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	Synchronous motors , Method of starting of synchronous motor.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Discuss the methods of starting of synchronous motors.
3	Tools used	Chalk, Blackboard and Projector
4	Lecture Description	**Methods of starting of synchronous motor** **The average synchronous motor torque is zero at the start. For a net average torque, it is necessary to rotate the rotor at a speed very near to synchronous speed. This is possible through various methods in practice.** **The different methods used to start a synchronous motor are :** **Using Pony Motors :** **By using the small pony motors like a small induction motor, we can start the synchronous motor. This small induction motor is coupled to the rotor of the synchronous motor. The function of this induction motor is to bring the rotor of the synchronous motor to the synchronous speed.** **Once the rotor attains the synchronous speed the pony motor is dis-coupled from the rotor. The synchronous motor continues to rotate at synchronous speed, by supplying d.c. excitation to the rotor through the slip-rings. One should remember that the motor used as the pony motor must have less number of poles than the synchronous motor used.** **Using Small D.C. Machine :** **In the above method, we have a seen small induction motor to start the motor. Here we use d.c. motor instead of induction motor to bring the motor to synchronous motor.**

		Once the d.c. motor brings the rotor of the synchronous motor to synchronous speed. The motor starts acting as the d.c. generator and starts giving excitation to the field winding of the synchronous motor. **Using Damper Winding :** **When a 3-phase supply is given to the synchronous motor it fails to start. In order to make it start copper bars circuited at both ends (similar to the squirrel cage rotor of an induction motor) are placed on the rotor, these bars or winding are known as 'Damper Winding'.** **Now when the supply is given the field winding setups a rotating magnetic field. Due to the damper winding used, the rotor starts rotating as an induction motor i.e., less than the synchronous speed at starting. Once d.c. excitation is given to the field winding and the motor is then pulled into synchronism.** **The damper winding is used to start the motor and hence can be used for starting purposes only. Because once the rotor rotates at synchronous speed the relative motion between the damper winding and rotating magnetic will be equal, and hence induced emf and current will be zero. The damper winding will be out of the circuit**
5	Questions	1. State the methods of starting of synchronous motor.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

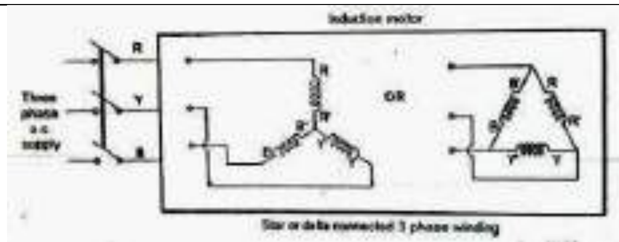
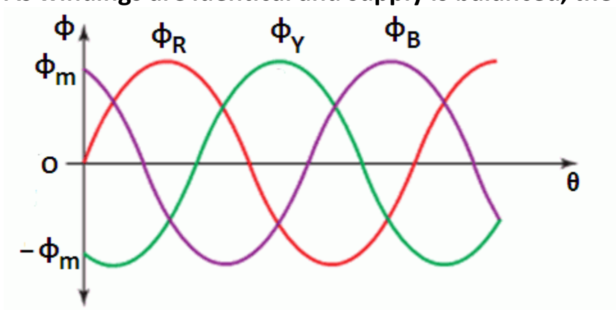
Lecture Plan-25

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	Synchronous motors , Applications of synchronous motors.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Say applications of synchronous motor
3	Tools used	Chalk, Blackboard and Projector
4	Lecture Description	**Applications of synchronous motor** **Power Factor Correction :** **Overexcited synchronous motors work as a synchronous condenser, are used for the improvement of power factor of industrial loads having lagging power factors.** **Constant Speed Applications :**

		Synchronous motor providing constant speed at high efficiency, so they can be used for constant speed applications such as centrifugal pumps, blowers, line shafts, motor-generator sets, synchronous clocks, air-compressors, textile mills, paper mills, cement mills, etc. **Improving Voltage Regulation of Long Transmission Lines :** **A synchronous motor with a field regulator can be used to control the voltage at the end of long transmission lines by varying its excitation.** **Frequency Changer :** **A synchronous motor can be used to drive another alternator to generate supply at different frequencies due to its constant speed nature. In this case, it is called a 'Frequency Changer'.**
5	Questions	1. State 3 applications of synchronous motor.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-26

Subject -Energy Conversion -|| Semester- 5^{th}		
1	Topic	3-φ Induction motor , Production of rotating magnetic field.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Explain production of rotating magnetic field of 3-φ induction motor.
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	**Production of rotating magnetic field** **The production of Rotating magnetic field in 3 phase supply is very interesting.** **When a 3-phase winding is energized from a 3-phase supply, a rotating magnetic field is produced. This field is such that its poles do no remain in a fixed position on the stator but go on shifting their positions around the stator. For this reason, it is called a rotating field.** **It can be shown that the magnitude of this rotating field is constant and is equal to 1.5 fm where fm is the maximum flux due to any phase.** **A three-phase induction motor consists of three phases winding as its stationary part called stator. The three-phase stator winding is connected in star or delta.** **The three-phase windings are displaced from each other by 120°. The windings are supplied by a balanced three phase ac supply.**

		 **The three-phase currents flow simultaneously through the windings and are displaced from each other by 120° electrical. Each alternating phase current produces its own flux which is sinusoidal.** **So all three fluxes are sinusoidal and are separated from each other by 120°.** **If the phase sequence of the windings is R-Y-B, then mathematical equations for the instantaneous values of the three fluxes ΦR , ΦY ,ΦB can be written as,** **$\Phi R = \Phi m \sin(\omega t)$** **$\Phi Y = \Phi m \sin(\omega t - 120)$** **$\Phi B = \Phi m \sin(\omega t - 240)$** **As windings are identical and supply is balanced, the magnitude of each flux is Φm .**  Fluxes produced by line currents
5	Questions	1. Describe the production of rotating magnetic field of 3-φ induction motor.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-27

Subject. -Energy Conversion -|| **Semester -5th**	

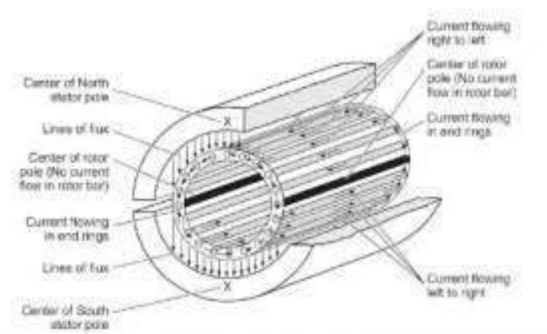
1	**Topic**	**CLASS TEST**
2	**Intended learning outcome**	
3	**Tools used**	**Chalk and blackboard**
4	**Problems**	
5	**Questions**	

Lecture Plan-28

<table>
<tr><td colspan="3">Subject -Energy Conversion -||
Semester- 5th</td></tr>
<tr><td>1</td><td>Topic</td><td>3-φ Induction motor , Constructional features of squirrel cage and slip ring induction motor.</td></tr>
<tr><td>2</td><td>Intended Learning Outcome</td><td>On successful completion of the Lecture students are able to:
1. Classify 3-φ induction motor.
2. Describe different parts of each induction motor.</td></tr>
<tr><td>3</td><td>Tools used</td><td>Chalk, Blackboard and Projector</td></tr>
<tr><td>4</td><td>Lecture Description</td><td>

Construction of squirrel cage induction motor

Parts that are required for the construction of squirrel cage induction motor are stator, rotor, fan, bearings. The stator consists of mechanically and electrically 120 degrees apart three-phase winding with metal housing and core. In order to provide the path of low reluctance for flux generated by AC current, the winding is mounted on the laminated iron core.



Motor Parts

Rotor converts given electrical energy into mechanical output. The shaft, a core, short-circuited copper bars are the parts of the rotor. In order to avoid hysteresis and eddy currents that are leading to power loss, the rotor is laminated. And I order to prevent cogging, conductors are skewed which also helps to give a good transformation ratio.

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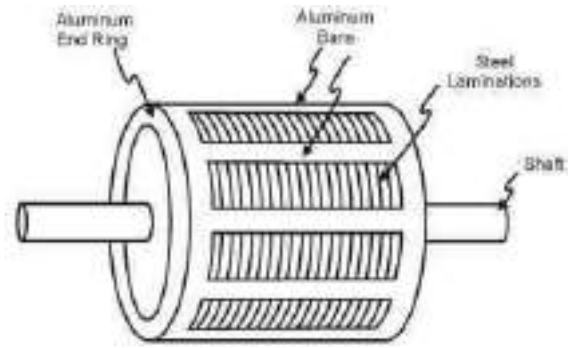
</table>

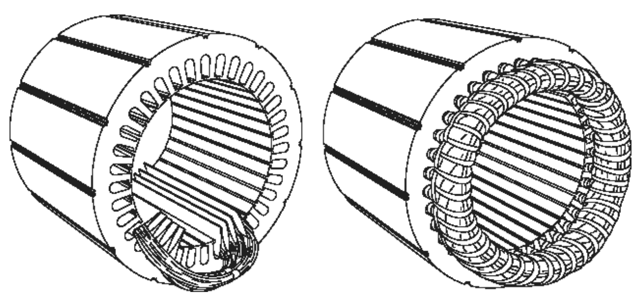


A fan attached at the back of the rotor for heat exchange helps in maintaining under a limit of the temperature of the motor. For the smooth rotation, bearings are provided in the motor.

Construction of Slip ring induction motor

A slip ring induction motor is an asynchronous motor, since the rotor never operates at synchronous speed with the stator poles. We will understand the construction and operation of the slip ring induction motor:

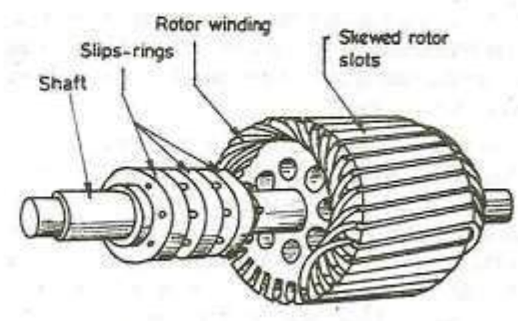
Stator:



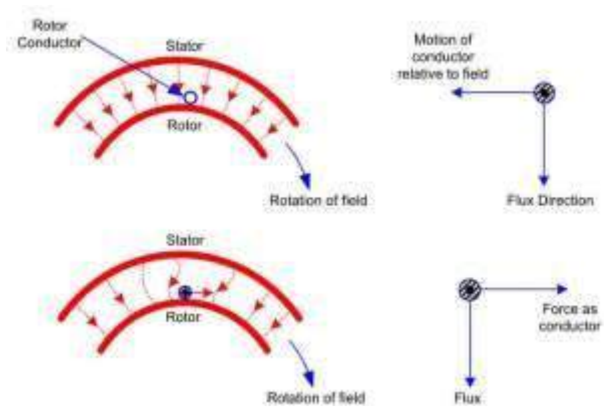
slip%20ring

The stator construction is the same for both the squirrel cage and the slip ring induction motor. The main difference in the slip ring induction motor is in the construction and use of the rotor. Some changes in the stator can be found when using a slip ring motor in a cascade system since the supply of the slave motor is controlled by the supply of the rotor of another slip ring motor with external resistance mounted on its rotor.

Rotor:

		 **Slip Ring Rotor** Slip ring induction motors usually have a “coiled phase” rotor. This type of rotor is provided with a distributed three-phase double-layer winding consisting of coils used in alternators. The rotor core is formed by steel laminates that have grooves to accommodate single-phase, 3-phase windings. These windings are placed at 120 degrees of electrical separation.
5	Questions	
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

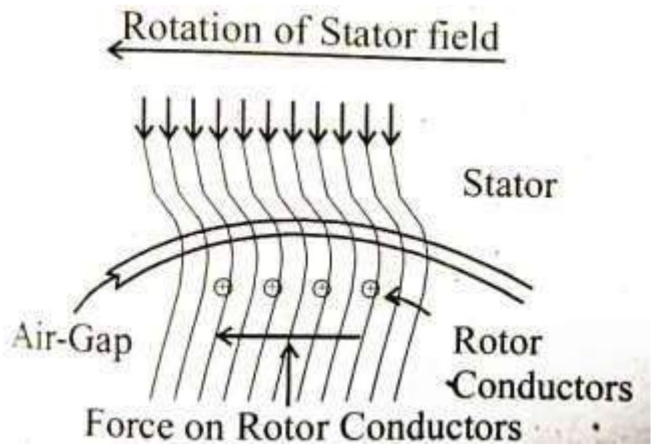
Lecture Plan-29

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	3-φ Induction motor , Working principle of operation of 3-φ induction motor
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Explain working principle of operation of 3-φ induction motor.
3	Tools used	Chalk, Blackboard and Projector
4	Lecture Description	For explaining the principle of operation of a three-phase induction motor, consider a portion of the three-phase induction motor as shown in the figure. 

working principle of 3-φ induction motor

The working of the three-phase induction motor is based on the principle of electromagnetic induction.

When three-phase stator winding of an induction motor is energized from a 3 phase supply, a rotating magnetic field is set up which rotates around the stator at synchronous speed (Ns).



A portion of rotating magnetic field in a three-phase induction motor

Synchronous Speed,

$$N_s = 120\, f/P$$

Where,

f = frequency

P = Number of Poles

This rotating field passes through the air gap and cuts the rotor conductors, which are stationary.

An EMF gets induced in every rotor conductor due to the relative speed between the rotating magnetic flux and the stationary rotor. Since the rotor circuit is short-circuited, currents start flowing in the rotor conductors.

The current-carrying rotor conductors are placed in the magnetic field produced by the stator. Consequently, a mechanical force acts on the rotor conductors. The sum of the

		mechanical forces on all the rotor conductors produces a torque which tends to move the rotor in the same direction as the rotating field. The fact that the rotor is urged to follow the stator field (i.e., rotor moves in the direction of stator field) can be explained by Lenz's law. According to Lenz's law, the direction of rotor currents will be such that they tend to oppose the cause of producing them. Now, the cause producing the rotor currents is the relative speed between the rotating field and the stationary rotor conductors. Hence to reduce this relative speed, the rotor starts running in the same direction as that of the stator field and tries to catch it. This is how a three-phase induction motor starts runningPhase Induction Motor.
5	Questions	State principle of operation of 3-φ induction motor , explain it's working with diagram.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-30

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	3-φ Induction motor , Slip speed ,slip & establish the relation of slip with rotor quantities
2	Intended Learning Outcome	On successful completion of the Lecture students are able to:. 1.Define slip speed & slip. 2.Explain relation of slip with rotor quantities.
3	Tools used	Chalk, Blackboard and Projector

4	Lecture Description	**Slip** In practice, the rotor never succeeds in ' catching up ' with the stator field.If it really did so , then there would be no relative speed between the two , hence no rotor e.m.f.,no rotor current & so no torque to maintain rotation.That is why the rotor runs at a speed which is always less than the speed of the stator field. The difference in speeds depends upon the load on the motor. The difference between the synchronous speed Ns & the actual speed N of the rotor is known as slip. Though it may be expressed in so many revolution/second , yet it is usual to express it as a percentage of the synchronous speed. Actually , the term 'slip ' is descriptive of the way in which the rotor 'slip back' from synchronism. % slip $s = N_s - N / N \times 100$ Sometimes, $N_s - N$ is called the slip speed. Obviously , rotor speed is $N = N_s(1-s)$ **Relation of slip with rotor quantities**
5	Questions	1. What do you mean by slip speed ? 2. Define slip. 3. Derive an expression for relation of slip with rotor quantities.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-31

Subject -Energy Conversion -||

Semester- 5th		
1	Topic	3-φ Induction motor , Expression for starting torque & running torque , & condition for maximum torque.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Define starting torque. 2. Define running torque. 3. State the condition for maximum torque.
3	Tools used	Chalk, Blackboard and Projector
4	Lecture Description	(see below)

Starting torque

The torque developed by the motor at the instant of starting is called starting torque. In some cases , it is greater than the normal running torque , whereas in some other cases it is somewhat less.

Let ,. $E2$ = rotor e.m.f. per phase at stand still ;

$R2$ = rotor resistance per phase

$X2$ =rotor reactance per phase at standstill

$X2 = \sqrt{(R2^2 + X2^2)}$ =rotor impedence per phase at standstill

Then ,

$I2 = E2/Z2 = E2/\sqrt{(R2^2 + X2^2)}$;

$\cos\varphi2 = R2/Z2 = R2/\sqrt{(R2^2 + X2^2)}$

Standstill or starting torque $Tst = k1\, E2\, I2 \cos\varphi2$

Or. $Tst = k1\, E2 \,.\, E2/\sqrt{(R2^2 + X2^2)} \times R2/\sqrt{(R2^2 + X2^2)}$

$= k1 E2^2 R2/R2^2 + X2^2$

If supply voltage V is constant , then the flux φ & hence , E2 both are constant .

:. $Tst = k2\, R2 / R2^2 + X2^2$

$= k2R2/Z2^2$, where k2 is some other constant.

Now ,. $k1 = 3/2\pi Ns$,

:. $Tst = 3/2\pi Ns \times E2^2 R2/R2^2 + X2^2$

Where Ns —> synchronous speed in rps.

Condition for maximum starting torque

It can be proved that starting torque is maximum when rotor resistance is equal to rotor reactance.

Now $T_{st} = k2R2 / R2^2 + X2^2$

$:.\ dT_{st} / dR2 = k2\ [1/R2^2+X2^2 - 2R2^2/(R2^2+X2^2)^2] = 0$

Or. $R2^2+X2^2=2R2^2$. $:.\ R2=X2$

Running torque

T proportional to ErIr cosφ2 or φIr cosφ2

Where Er = rotor e.m.f. /phase under running conditions

Ir = rotor current per phase under running conditions

Now $Er = sE2$

:. $Ir = Er / Zr = sE2/\sqrt{R2^2 + (sX2)^2}$

$\cos \varphi 2 = R2/\sqrt{R2^2 + (sX2)^2}$

:. T proportional to $s\ \varphi\ E2\ R2 / R2^2+sX2^2 = k\ \varphi.s.E2R2 / R2^2+ (sX2)^2$

Also. $T = k1\ sE2^2\ R2/ R2^2+(sX2)^2$

Where k1 is another constant . It's value can be proved to be equal to

3/2 πNs. Hence in that case , expression for torque becomes

$$T = 3/2\pi Ns \times sE2^2R2/R2^2 +(sX2)^2 = 3/2\pi Ns \times sE2^2R2 /Zr^2$$

At standstill when s=1 , obviously

$$T_{st} = k1E2^2R2/R2^2+X2^20$$

Condition for maximum running torque

The torque of a rotor under running conditions is

$$T = k\ \varphi\ sE2R2/R2^2 + (sX2)^2$$

$$= k1\ sE2^2\ R2 / R2^2 + (sX2)^2$$

		The condition for maximum torque may be obtained by differentiating the above expression with respect to slip s and then putting it equal to zero. However , it is simpler to put Y = 1/T & then differerentiate it. :. $Y = R2^2 + (sX2)^2 / k\,\varphi\, sE2R2 = R2/k\,\varphi sE2 + sX2^2$; $dY/ds = -R2/k\,\varphi s^2\, E2 + X2^2/k\,\varphi E2R2 = 0$:. $R2/\,k\,\varphi s^2 E2 = X2^2/k\,\varphi E2R2$ or ,. $R2^2 = s^2 X2^2$ or , $R2 = sX2$ Hence , torque under running condition is maximum at that value of the slip s which makes rotor reactance per phase equal to rotor resistance per phase. The slip is sometimes written as Sb and the maximum torque is Tb . Slip corresponding to maximum torque is $s = R2/X2$ Putting $R2 = sX2$ in the above equation for the torque , we get $Tmax = k\,\varphi s^2 E2X2/2s^2\, X2^2$ or,. $Tmax = KE2/2\,X2$ Substituting value of $s = R2/X2$ in the 1st equation above , we get $Tmax = k1\,(R2/X2)\,E2^2.R2/R2^2 + (R2/X2)^2.X2^2$ $= k1E2^2/2X2$ Since , $k1 = 3/2\pi Ns$, we have , $Tmax = 3/2\pi Ns \times E2^2/2X2$ N-m
5	Questions	1. Derive an expression for starting torque. 2. Derive an expression for running torque. 3. Derive an expression for maximum torque.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-32

UNIT DISCUSSION-II

1. What are the types of alternator ?

2. What frequency voltage must be applied to a 20-pole synchronous motor so that it runs at a speed of 300 rpm ?

3. Can a three phase IM never runs at synchronous speed ?

4. How frequency of rotor current is related to slip ?

5. Why are alternator rated in kVA , not in kW ?

6. Why single phase induction motor is not self starting ?

7. If a single phase induction motor fails to start but gives a humming sound , what is the reason behind it ?

8. What are the two advantages of short pitched coil in an alternator wdg.

9. Calculate the value of 'Kd' for a 36-slots , 4-poles single layer three phase alternator.

10. For traction purpose which motors are used ?

11. State two applications of DC series motor.

12. State the difference between coil pitch and pole pitch.

13. State pitch factor and distribution factor.

14. What is the function of damper bars ?

15. What is the principle of operation of synchronous motor ?

16. Why a starting winding is needed ?

17. In which rotor high starting torque is produced and why ?

18. What are the types of alternator ?

19. What frequency voltage must be applied

20. Why rotor slots of an induction motor are skewed ?

21. What is plugging ?

23. Write down the advantages of stationary armature of alternator ?

24. What do you mean by voltage regulation of an alternator ?

25. What is damper bar or damper winding ?

26. State the applications of synchronous motor.

27. How the direction of rotation of a single phase induction motor can be reversed ?

28. Why parallel operation of alternators needed ?

29. What is hunting ?

30. What is V-curves of synchronous motor ?

31. What do you mean by infinite bus ?

32. Which type of alternator is used in hydroelectric power plants and why ?

33. What is the relation between electrical degrees and mechanical angle of an alternator ?

34. State the difference between coil pitch and pole pitch.

35. State pitch factor and distribution factor.

36. What is the function of damper bars ?

37. What is the principle of operation of synchronous motor ?

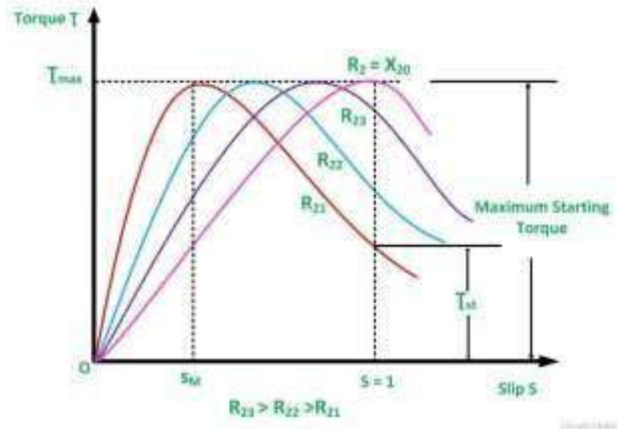
38. Why a starting winding is needed ?

39. In which rotor high starting torque is produced and why ?

S.K.D.A.V Govt. Polytechnic, Rourkela

Lecture Plan-33

Subjec. -Energy Conversion -|| Semester- 5th		
1	Topic	3-φ Induction motor , Torque slip characteristics
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Define torque. 2. Discuss torque slip characteristics.
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	**Torque Slip Characteristics of Three Phase Induction Motor** **The torque slip curve for an induction motor gives us the information about the variation of torque with the slip. The slip is defined as the ratio of difference of synchronous speed and actual rotor speed to the synchronous speed of the machine. The variation of slip can be obtained with the variation of speed that is when speed varies the slip will also vary and the torque corresponding to that speed will also vary.**



The curve can be described in three modes of operation-

torque slip curve for three phase induction motor

The torque-slip characteristic curve can be divided roughly into three regions:

Low slip region

Medium slip region

High slip region

Motoring Mode

In this mode of operation, supply is given to the stator sides and the motor always rotates below the synchronous speed. The induction motor torque varies from zero to full load torque as the slip varies. The slip varies from zero to one. It is zero at no load and one at standstill. From the curve it is seen that the torque is directly proportional to the slip.

That is, more is the slip, more will be the torque produced and vice-versa. The linear relationship simplifies the calculation of motor parameter to great extent.

Generating Mode

In this mode of operation induction motor runs above the synchronous speed and it should be driven by a prime mover. The stator winding is connected to a three phase supply in which it supplies electrical energy. Actually, in this case, the torque and slip both are negative so the motor receives mechanical energy and delivers electrical energy. Induction motor is not much used as generator because it requires reactive power for its operation.

That is, reactive power should be supplied from outside and if it runs below the synchronous speed by any means, it consumes electrical energy rather than giving it at the output. So, as far as possible, induction generators are generally avoided.

Braking Mode

In the Braking mode, the two leads or the polarity of the supply voltage is changed so that the motor starts to rotate in the reverse direction and as a result the motor stops. This

		method of braking is known as plugging. This method is used when it is required to stop the motor within a very short period of time. The kinetic energy stored in the revolving load is dissipated as heat. Also, motor is still receiving power from the stator which is also dissipated as heat. So as a result of which motor develops enormous heat energy. For this stator is disconnected from the supply before motor enters the braking mode. **If load which the motor drives accelerates the motor in the same direction as the motor is rotating, the speed of the motor may increase more than synchronous speed. In this case, it acts as an induction generator which supplies electrical energy to the mains which tends to slow down the motor to its synchronous speed, in this case the motor stops. This type of breaking principle is called dynamic or regenerative breaking.**
5	Questions	1. What is the relation between slip and torque.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-34

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	3-φ Induction motor , Relation between full load torque and starting torque.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Discuss the relation between full load torque and starting torque.
3	Tools used	Chalk, Blackboard
4	Lecture Description	**Relation between full load torque and starting torque** Let sf be the slip corresponding to full load torque , then Tf proportional to $sfR2/R2^2+(sfX2)^2$ and , Tmax proportional to $1/2\times X^2$:. Tf /Tmax = $2sf\ R2X2/R2^2+(sf\ X2)^2$ Dividing both the numerator and denominator by $X2^2$, we get :. Tf /Tmax = $2sf. (R2/X2)/(R2/X2)^2+sf^2 = 2asf/a^2+sf^2$ Where a = R2/X2=resistance/standstill reactance. In general , operating torque at any slip s/maximum torque $=2as/a^2+sf^2$ Tst proportional to $R2\ /R2^2+X2^2$ Tmax proportional to 1/2X2 :. Tst/Tmax = $2R2X2/R2^2+X2^2 =2(R2/X2)/1+ (R2/X2)^2 =2a/1+a^2$ Where a = R2/X2 =rotor resistance/standstill reactance per phase.

		Hence ,. $Tf/Tst = sf$ Or,. $Tf = sf\ Tst$
5	Questions	1. Derive the relation between full load torque and starting torque. 2. What do you mean by maximum torque.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-35

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	3-φ Induction motor , Relations between rotor copper loss , rotor output & gross torque & relation of slip with rotor copper loss.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Relate rotor copper loss , rotor output & gross torque. 2. Relate slip with rotor copper loss.
3	Tools used	Chalk, Blackboard and Projector
4	Lecture Description	Stator input. P1 = stator output + stator losses The stator output is transferred entirely inductively to the rotor circuit. Obviously , rotor input ,. P2 = stator output Rotor gross output,. Pm = rotor input P2 − rotor Cu losses This rotor output is converted into mechanical energy and gives rise to gross torque Tg. Out of these gross torque developed , some is lost due to windage and friction losses in the rotor and the rest appears as the useful and shaft torque Tsh. Let N r.p.s. be the actual speed of the rotor and if Tg is in N-m , then $Tg \times 2\pi N$ = rotor gross output in watts , Pm :. Tg = rotor gross output in watts , Pm /2πN. N-m .... (1) If there were no Cu losses in the rotor , then rotor output will equal rotor input and the rotor will run at synchronous speed. :. Tg = rotor input , P2/2πNs...............................(2) From 1 & 2 , we get ,

		Rotor gross output. $P_m = T_g \times 2\pi N$ Rotor Input $P_2 = T_g \times 2\pi N_s$. ....(3) The difference of two equals rotor Cu loss . :. Rotor Cu loss $= P_2 - P_m = T_g \times 2\pi (N_s - N)$. ....(4) From 3 & 4,. rotor Cu loss / rotor input = $N_s - N$. $/N_s = s$:. Rotor Cu loss = s× rotor input = s× power across air gap = sP_2 Also, rotor input =. rotor Cu loss / s Rotor gross output , P_m = input P_2 − rotor Cu loss = input −s× rotor input $= (1-s)$ input P_2 :. Rotor gross output $P_m = (1-s)$ rotor input P_2 or ,. $P_m/P_2 = 1-s = N/N_s$ So , $P_m / P_2 = N/N_s$:. rotor efficiency = N/N_s Also , rotor Cu loss / rotor gross output = s/(s−1)
5	Questions	1. Establish the relation among rotor Cu loss , rotor output & gross torque. 2. Establish the relation between slip and rotor copper loss.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-37

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	3-φ Induction motor , method of starting & different types starters used in 3-φ induction motor.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Say methods of starting of 3-φ induction motor. 2. Classify starters.
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	**Methods of starting and Different types of starters used in induction motor** A three-phase Induction Motor is Self Starting. When the supply is connected to the stator of a three-phase induction motor, a rotating magnetic field is produced, and the rotor

begins rotating and the induction motor starts. At the time of starting, the motor slip is unity, and the starting current is very large.

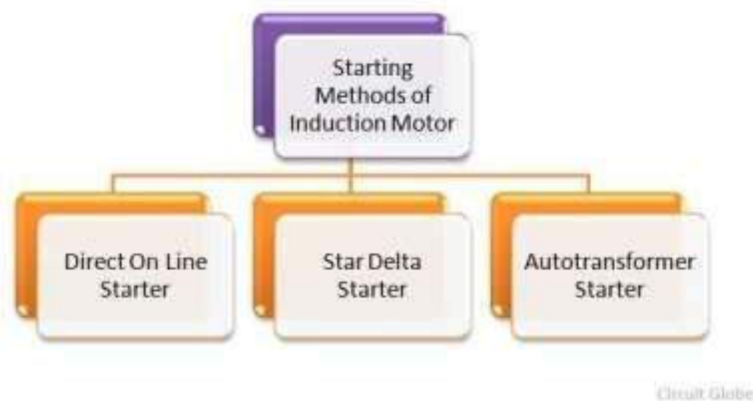
The purpose of a starter is not to just start the motor, but it performs the two main functions. They are as follows:

To reduce the heavy starting current,

To provide overload and under-voltage protection.

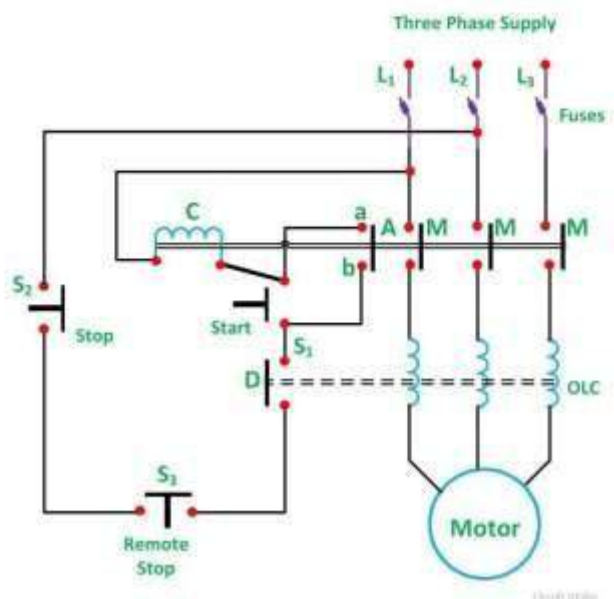
The three-phase induction motor may be started by connecting the motor directly to the full voltage of the supply. The motor can also be started by applying a reduced voltage to the motor. The torque of the induction motor is proportional to the square of the applied voltage. Thus, greater torque is exerted by a motor when it is started on full voltage than when it is started on the reduced voltage.

There are three main methods of Starting of Cage Induction Motor. They are as follows:



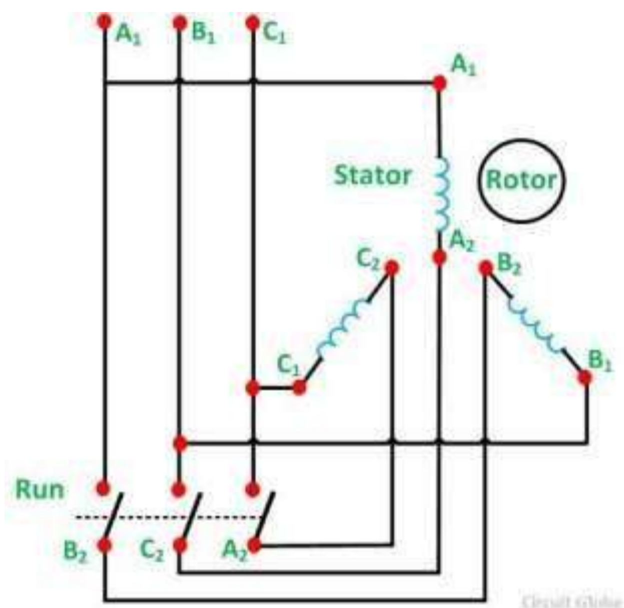
Direct on-line starter

The direct on-line starter method, of an induction motor, is simple and economical. In this method, the starter is connected directly to supply voltage. By this method, small motors up to 5 kW rating are started to avoid the supply voltage fluctuation.



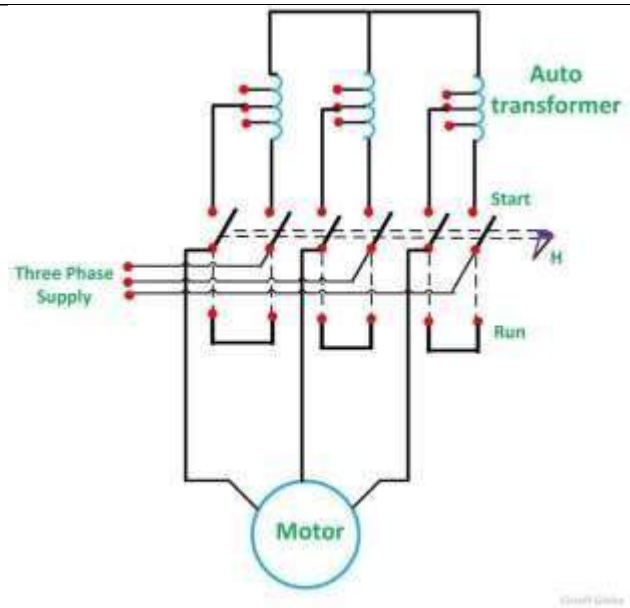
Star delta starter

The star delta starter method of starting three-phase induction motors is very common and widely used among all the methods. In this method, the motor runs at delta-connected stator windings.



Autotransformer starter

The Autotransformer is used in both types of connections, i.e., either star connected or delta connected. The autotransformer is used to limit the starting current of the induction motor.

		 The above three starters are used for the cage rotor induction motor.
5	Questions	1. What is the function of starter ? 2. Discuss types of starters used for induction motor.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-38

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	3-φ Induction motor , Speed control by voltage control , rotor resistance control , pole changing & frequency control method.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Define speed control of an induction motor. 2. State different methods for speed control.
3	Tools used	Chalk, Blackboard and Projector

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<td>Lecture Description</td>
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Speed control by voltage control

This method , though the cheapest and easiest , is rarely used because

1. A large change in voltage is required for a relatively small change in speed.

2. This change in voltage will result in a large change in the flux density thereby seriously disturbing the magnetic conditions of the motor.

Speed control by rotor resistance control

In this method , which is applicable to slip ring motors alone , the motor speed is reduced by introducing an external resistance in the rotor circuit. For this purpose , the rotor starter may be used , provided it is continuously rated. This method is , in fact , similar to the armature rheostat control method of d.c. shunt motors.

It has been shown in figure , that near synchronous speed ,T proportional to s/R2.

It is obvious that for a given torque , slip can be decreased by increasing the rotor resistance R2.

One serious disadvantage of this method is that with increase in rotor resistance , I^2R losses also increase which decrease the operating efficiency of the motor. In fact , the loss is directly proportional to the reduction in the speed.

The second disadvantage is the double dependence of speed , not only on R2 but on load as well. Because of the wastefulness of this method , it is used where speed speed changes are needed for short periods only.

Speed control by pole changing

This method is easily applicable to squirrel cage motors because the squirrel cage rotor adopts itself to any reasonable number of stator poles.

Synchronous speed of an induction motor could also be changed by changing the number of stator poles. This change of number of poles is achieved by having two or more entirely independent stator windings in the same slots. Each winding gives a different no. of poles and hence different synchronous speed. For example , a 36 slot stator may have two 3-φ windings , one with 4 poles and the other with 6-poles. With a supply frequency of 50 Hz , 4-pole winding will give Ns = 120×50/4=1500rpm and the 6 pole winding will give Ns =120×50/6=1000rpm. Motors with four independent stator winding are also in use and they give 4 different synchronous speeds. Of course , one winding is used at a time , the others being entirely disconnected.

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This method has been used for elevator motors , traction motors and also for small motors driving machine tools.

Speeds in the ratio of 2:1 can be produced by a single winding if wound on the consequent pole principle. In that case , each of the two stator windings can be connected by a simple switch to give two speeds , each , which means four speeds in all. For example , one stator winding may give 4 or 8 poles and the other 6 or 12 poles. For a supply frequency of 50 Hz , the four speeds will be 1500 , 750 , 1000 & 500 rpm. Another combination , commonly used , is to group 2 & 4-pole winding with a 6 & 12-pole winding , which gives four synchronous speeds of 3000 ,1500 , 1000 & 500 rpm.

Speed control by frequency control

This method is also used very rarely. We have seen that the synchronous speed of an induction motor is given by $N_s = 120f/P$. Clearly , the synchronous speed of an induction motor can be changed by changing the supply frequency f. However , this method could only be used in cases where the induction motor happens to be the only load on the generators , in which case , the supply frequency could be controlled by controlling the speed of the prime movers of the generators. But , here again the range over which the motor speed may be varied is limited by the economical speeds of the prime movers. This method has been used to some extent on electrically driven ships.

5	Questions	1. Explain Speed control by voltage control. 2. Explain speed control by rotor resistance control. 3. Explain speed control by pole changing. 4. Explain speed control by frequency control.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-39

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	3-φ Induction motor , Plugging as applicable to 3-φ induction motor , different types of motor enclosures
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Say plugging of 3-φ induction motor. 2. Explain the types of motor enclosures.
3	Tools used	Chalk, Blackboard & Projector
4	Lecture Description	**Plugging as applicable** An induction motor can be quickly stopped by simply interchanging any of its two stator leads. It reverses the direction of the revolving flux which produces a torque in the reverse direction , thus applying brake on the motor. Obviously , during this so-called plugging period , the motor acts as a brake. It absorbs kinetic energy from the still revolving load causing it's speed to fall. The associated power Pm is dissipated as heat in the rotor. At the same time , the rotor also continues to receive power P2 from the stator which is also dissipated as heat. Consequently , plugging produces rotor I^2R losses which even exceed those when the rotor is locked. **Motor enclosures** Enclosed and semi enclosed motors are practically identical with open motors in mechanical construction and in their operating characteristics.

Different types of motor enclosures

Many different types of frames or enclosures are available to suit particular requirements. Some of the common type enclosures are described below :

1. Totally enclosed , Non ventilated type

Such motors have solid frames and end shields , but no openings for ventilation. They get cooled by surface radiation only. Such surface cooled motors are seldom furnished in sizes above two or three kW , because higher ratings require frames of much larger sizes than fan cooled motors of corresponding ratings.

2.Splash proof type

In the frames of such motors , the ventilating openings are so constructed that the liquid drops or dust particles falling on the motor are coming towrds it in a straight line at any angle not agreater than 100° from the vertical are not able to enter the motor either directly or by striking and running along the.

3. Totally enclosed , Fan cooled type

In such motors , cooling air is drawn into the motor by a fan mounted on the shaft. This air is forced through the motor between the inner fully enclosed frame and an outer shell , over the end balls and the stator laminations and is then discharged through openings in the opposite side.

An internal fan carries the generated heat to the totally enclosed frame , from where it is conducted to the outside. Because of totally enclosing frame , all working parts are protected against corrosive or abrasive effects of fumes , dust , and moisture.

4. Cowl covered motor

These motors are simplified form of fan cooled motors. These consist of totally enclosed frame with a fan and cowl mounted at the end opposite to the driving end. The air is drawn into the cowl with the help of fan and is then forced over the frame. The contours of the cowl guide the cooling air in proper directions. These motors are superior to the usual fan cooled motors for operation in extremely dusty atmosphere i.e. gas works , chemical works , collieries and quarries etc. because their are no air passages which will become clogged with dust.

5. Protected type

This construction consists of perforated covers for the openings in both end shields.

6. Drip proof motors

The frames of such motors are so constructed that liquid drops or dust particles , falling on the machine at any greater than 15° from the vertical , cannot enter the motor ,either directly or by striking and running along a horizontal or inwardly inclined smooth surface.

7.Self ventilated type

The construction of such motors consists of enclosed shields with provision for pipe connection on both the shields. The motor fan circulates sufficient air through pipes which are of ample section.

		8. Separately ventilated type These motors are similar to the self ventilated type except that ventilation is provided by a separated blower.
5	Questions	1. How 3-φ induction motor work as plugging. 2. What are the different types of motor enclosures.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-40

Subjet. -Energy Conversion -|| Semester- 5th		
1	Topic	**CLASS TEST QUESTION DISCUSSION AND DISTRIBUTION OF EVALUATED ANSWER SHEET TO THE STUDENT**
2	Intended Learning Outcome	On successful completion of the Lecture students are able to:
3	Tools used	Chalk, Blackboard
4	Lecture Description	
	Problems-	
5	Questions	
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-41

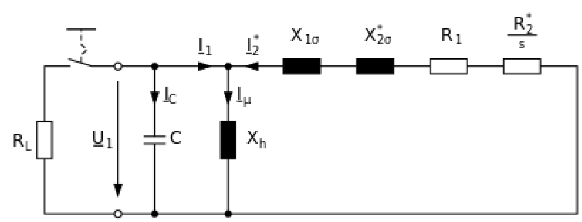
Subject -Energy Conversion -|| Semester- 5th		
1	Topic	3-φ Induction motor , Principle of induction generator & it's applications
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Say basic principle of induction generators. 2. Where these are used?
3	Tools used	Chalk, Blackboard
4	Lecture Description	Principle of induction generators

An induction generator produces electrical power when its rotor is turned faster than the synchronous speed. For a typical four-pole motor (two pairs of poles on stator) operating on a 60 Hz electrical grid, the synchronous speed is 1800 rotations per minute (rpm). The same four-pole motor operating on a 50 Hz grid will have a synchronous speed of 1500 RPM. The motor normally turns slightly slower than the synchronous speed; the difference between synchronous and operating speed is called "slip" and is usually expressed as per cent of the synchronous speed. For example, a motor operating at 1450 RPM that has a synchronous speed of 1500 RPM is running at a slip of +3.3%.

In normal motor operation, the stator flux rotation is faster than the rotor rotation. This causes the stator flux to induce rotor currents, which create a rotor flux with magnetic polarity opposite to stator. In this way, the rotor is dragged along behind stator flux, with the currents in the rotor induced at the slip frequency.

In generator operation, a prime mover (turbine or engine) drives the rotor above the synchronous speed (negative slip). The stator flux still induces currents in the rotor, but since the opposing rotor flux is now cutting the stator coils, an active current is produced in stator coils and the motor now operates as a generator, sending power back to the electrical grid.

Excitation



Equivalent circuit of induction generator

An induction generator requires an externally-supplied armature current. Because the rotor field always lags behind the stator field, the induction generator always consumes reactive power, regardless of whether it is operating as a generator or a motor.

A source of excitation current for magnetizing flux (reactive power) for the stator is still required, to induce rotor current. This can be supplied from the electrical grid or, once it starts producing power, from the generator itself. The generating mode for induction motors is complicated by the need to excite the rotor, which begins with only residual magnetization. In some cases, that residual magnetization is enough to self-excite the motor under load. Therefore, it is necessary to either snap the motor and connect it momentarily to a live grid or to add capacitors charged initially by residual magnetism and providing the required reactive power during operation. Similar is the operation of the induction motor in parallel with a synchronous motor serving as a power factor compensator. A feature in the generator mode in parallel to the grid is that the rotor speed is higher than in the driving mode. Then active energy is being given to the grid.[1]Another disadvantage of induction motor generator is that it consumes a significant magnetizing current I0 = (20-35)%.

An induction generator can be started by charging the capacitors, with a DC source, while the generator is turning typically at or above generating speeds. Once the DC source is removed the capacitors will provide the magnetization current required to begin producing voltage.

		An induction generator that has recently been operating may also spontaneously produce voltage and current due to residual magnetism left in the core. **Active power** Active power delivered to the line is proportional to slip above the synchronous speed. Full rated power of the generator is reached at very small slip values (motor dependent, typically 3%). At synchronous speed of 1800 rpm, generator will produce no power. When the driving speed is increased to 1860 rpm (typical example), full output power is produced. If the prime mover is unable to produce enough power to fully drive the generator, speed will remain somewhere between 1800 and 1860 rpm range. **Required capacitance** A capacitor bank must supply reactive power to the motor when used in stand-alone mode. The reactive power supplied should be equal or greater than the reactive power that the generator normally draws when operating as a motor. **Torque vs. slip** The basic fundamental of induction generators is the conversion from mechanical energy to electrical energy. This requires an external torque applied to the rotor to turn it faster than the synchronous speed. However, indefinitely increasing torque doesn't lead to an indefinite increase in power generation. The rotating magnetic field torque excited from the armature works to counter the motion of the rotor and prevent over speed because of induced motion in the opposite direction. As the speed of the motor increases the counter torque reaches a max value of torque (breakdown torque) that it can operate until before the operating conditions become unstable. Ideally, induction generators work best in the stable region between the no-load condition and maximum torque region. **Rated current** The maximum power that can be produced by an induction motor operated as a generator is limited by the rated current of the generator's windings.
5	Questions	1. Explain how induction motor operating as a induction generator. 2. Write the applications of induction generators.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-42

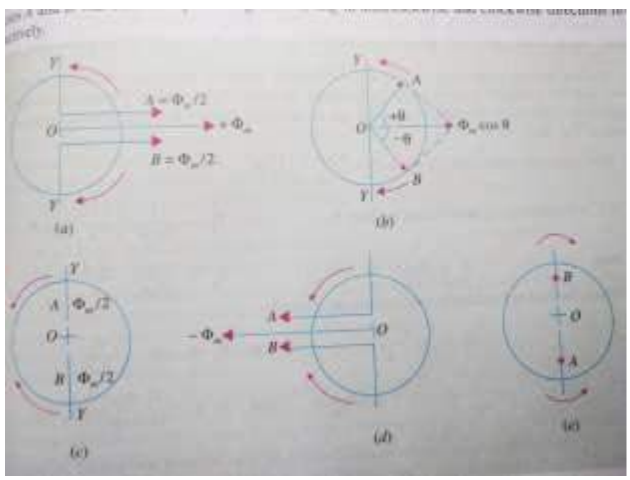
Subject -Energy Conversion -|| Semester- 5th		
1	Topic	1-φ Induction motor , Ferrari's principle , double revolving field theory & cross field theory to analyze starting torque of 1-φ induction motor.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. State Ferrari's principle. 2. Define double revolving field theory.

		3. Define cross field theory.
3	Tools used	Chalk, Blackboard & Projector
4	Lecture Description	

Ferrari's principle or double revolving field theory

This theory makes use of the idea that an alternating uni axial quantity can be represented by two oppositely rotating vectors of half magnitude. Accordingly , an alternating sinosoidal flux can be represented by two revolving fluxes , each equal to half of the value of the alternating flux and each rotating synchronously in opposite direction.

As shown in figure (a) , let the alternating flux have a maximum value of φm . It's component fluxes A and B will each be equal to $\varphi m/2$ revolving in anticlockwise and clockwise direction respectively.

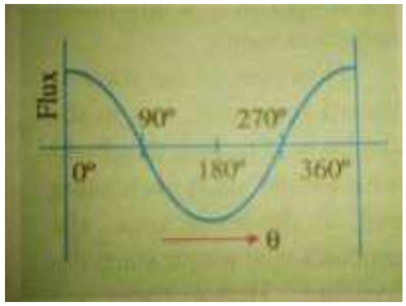


After sometime , when A and B would have rotated through angle θ and $-\theta$, as shown in figure (b) , the resultant flux would be

$$= 2\times \varphi m /2 \cos 2\theta/2 = \varphi m \cos\theta$$

After a quarter cycle of rotation , fluxes A and B will be oppositely directed as shown in figure (c) , so that the resultant flux will be zero.

After a half cycle , fluxes A and B will have a resultant of $-2\times\varphi m /2 = -\varphi m$. After three quarters of a cycle , again the resultant is zero as shown in figure (e) and so on. If we plot the values of resultant flux against θ between limits $\theta=0°$ to $\theta=360°$, then a curve similar to the one shown in figure below is obtained. That is why an alternating flux can be looked upon as composed of two revolving fluxes , each of half the value and revolving synchronously in opposite directions.



It may be noted that if the slip of the rotor is s with respect to the forward rotating flux , then its slip with respect to the backward rotating flux is 2−s.

Each of the two component fluxes , while revolving around the stator , cuts the rotor , induces an emf and this produces its own torque. Obviously , the two torques are oppositely directed , so that the net or resoltant torques is equal to their difference as shown in figure below.

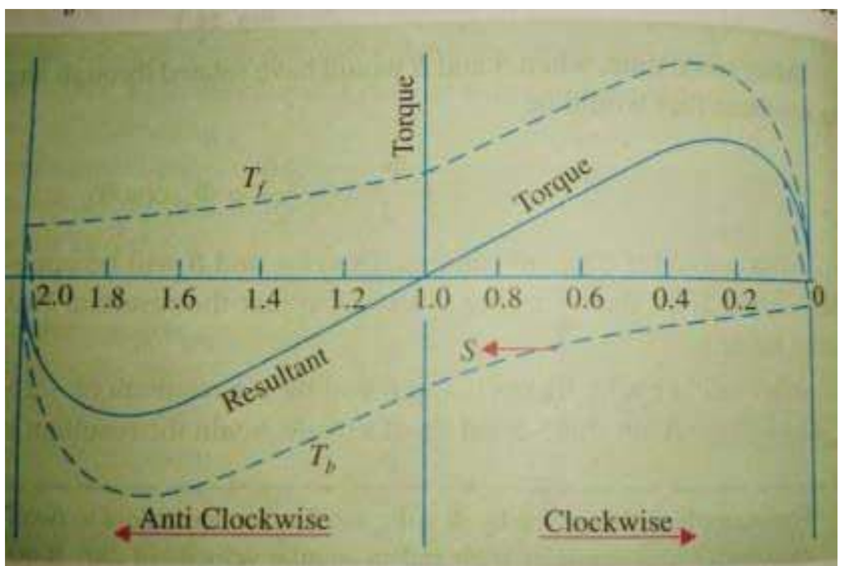


Fig. 36.3

Now , power developed by a rotor is Pg=(1-s)/s ×I2^2R2

If N is the rotor r.p.s. , then torque is given by

Tg= 1/2πN . [(1-s)/s] I2^2R2

Now , N = Ns(1-s)

:. Tg =1/2πNs . I2^2R2/s =k.I2^2R2/s

Hence , the forward and backward torques are given by

$Tf = K.I2^{2}R2/s$.   And , $Tb = -K.I2^{2}R2/(2-s)$

Or ,. $Tf = I2^{2}R2/s$ synch. watt , and , $Tb = -I2^{2}R2/2\text{-}s$ synch. watt

Total torque $T=Tf +Tb$

The above figure shows both torques and the resultant torque for slips between zero and +2. At standstill , s=1 , and 2−s= 1. Hence , Tf and Tb are numerically equal but , being oppositely directed , produce no resultant torque. That explains why there is no starting torque in a single phase induction motor.

However , if the rotor is started somehow , say , in the clockwise direction , the clockwise torque starts inreasing and , at the same time , the anticlockwise torque starts decreasing. Hence , there is a certain amount of net torque in the clockwise direction which accelerates the motor to full speed.

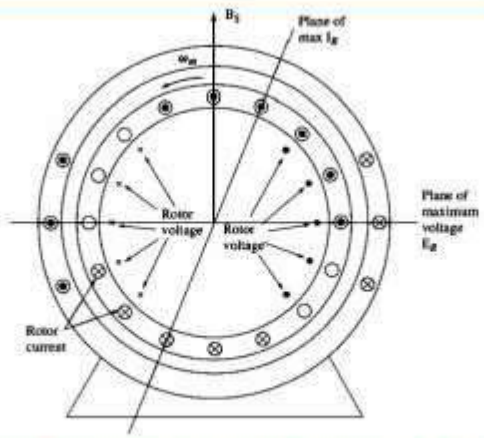
Cross field theory

The cross-field theory of single phase induction motor helps to view the induction motor with a different point of view.

This theory is related to the voltage and current which are static stator field can produce in the windings of the rotor when the rotor is rotating.

Let us assume that single-phase induction motor having a rotor that is taken to speed through the use of any exterior technique.

This type of motor can see here.



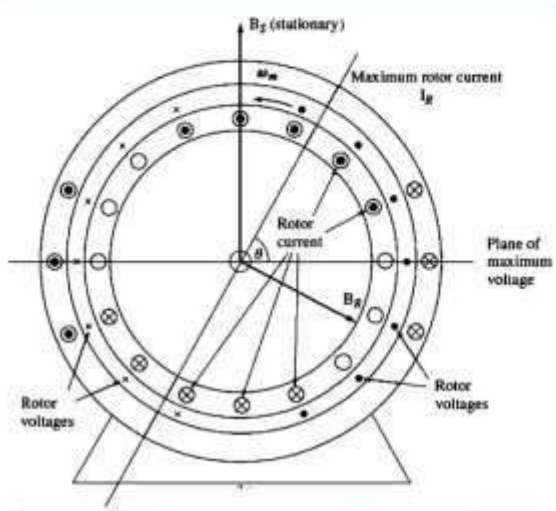
The development of induced torque in a single-phase induction motor, as explained by the crossfield theory. If the stator field is pulsing, it will induce voltages in the rotor bars, as shown by the marks inside the rotor. However, the rotor current is delayed by nearly 90° behind the rotor voltage, and if the rotor is turning, the rotor current will peak at an angle different from that of the rotor voltage.

Voltage is produced in the rotors through the peak voltage existing in the windings moving directly through the windings existing in the stator.

This rotor voltage generates a current in the rotor but the sine of the high value of reactance of the rotor, the current lags the voltage through the angle of ninety degrees.

The rotor rotating speed is almost equal to the synchronous speed so ninety degrees lagging factor in current generates the ninety degrees angular shift among the plane of peak rotor voltage and the plane of peak current.

In the below figure you can see the resultant rotor field.



This delayed rotor current produces a rotor magnetic field at an angle different from the angle of the stator magnetic fie ld.

		This delayed rotor current produces a rotor magoetic field at an angle different from the angle of the stator magnetic field. The rotor field is less than the stator field, since the loss of rotor but there is difference of ninety degrees angle in space and time as well. If these 2 fields are added for different time intervals so we can see that the net field in the motor is revolving in the anticlockwise direction. The vector sum of the rotor and stator magnetic fields at various times. showing a net magnetic field which rotates in a counterclockwise direction through the rotating field existing in the motor induction motor will create the total torque in the direction of movement and that torque will retain the rotor moving. If the rotor of the motor is moving in the clockwise direction the net torque will have clock-wise direction will again retain the motor moving.
5	Questions	1. Explain Cross field theory. 2. Explain double revolving field theory with diagram.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-43

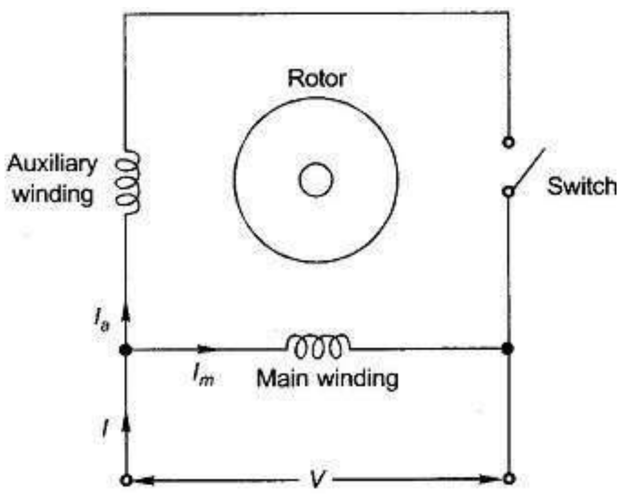
Subject -Energy Conversion -|| Semester- 5th		
1	Topic	1-φ Induction motor , Working principle , Torque characteristics , performance characteristics & application of following 1-φ induction motors: a. Split phase motor b. Capacitor start motor c. Capacitor start , capacitor run motor d. Permanent capacitor type motor e. Shaded pole motor
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Define split phase motor and capacitor start motor. 2. Explain working principle of different types of 1-φ induction motor.
3	Tools used	Chalk, Blackboard and PPT

4	Lecture Description	**Split phase motor**

Split phase motor

Working Principle

In split-phase motor two windings named main winding and starting winding are provided. At the time of starting, both the main and starting windings should be connected across the supply to produce the rotating magnetic field, and when the supply is given to the stator the rotating magnetic field is produced. The rotor is of a squirrel cage type and the revolving magnetic field sweeps part of the stationary rotor, inducing EMF in the rotor.



(a)

As the rotor bars are short-circuited, a current flows through them producing a magnetic field. This magnetic field opposes the revolving magnetic field and will combine with the main field to produce a revolving field. Once the rotor starts rotating and attains a speed up to 75 to 80 percent of the synchronous speed, the starting winding can be disconnected from the supply by the centrifugal switch

Performance and Characteristics

The starting torque is about twice the full load torque.

Speed falls with the increase in load with only about 5% to 7% otherwise it is a constant speed motor.

The current at the start is about 6 to 8 times.

Actual speed is less than synchronous speed Ns.

For the same weight, its rating is about 60 percent that of the polyphase motor. The split-phase induction motor has lower p.f. and lesser efficiency. Its P.f. is about 0.6 and efficiency is also about 60%.

It is suitable for easily starting loads where the frequency of starting is limited. This type of motor is not used for drives that require more than 1 KW because of the low starting torque.

Applications

The applications of this motor include different loads which are used for general purpose. Due to the excellent starting torque and easy direction-reversal characteristics, it finds use in lathe machines, drilling, washing machines, woodworking tools, belted fans, drill presses, oil burners, centrifugal pumps, compressors, air conditioning fans, floor polishers, blowers dryers, mixer grinder, heating blowers with belt-driven and conveyors with tiny belt-driven, and various other low starting torque applications.

This motor is used where the distribution of the three phases is not required.

This motor does not give lots of starting torque, thus the load should be quite small, and where mechanical gain can be used to assist the motor to begin.

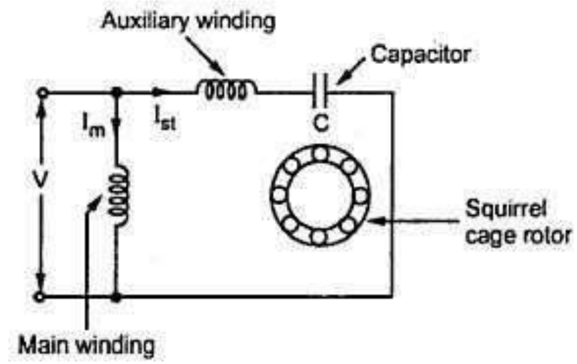
So, there you have every single fact to the question of what is split-phase induction motors. If you enjoyed this article in Linquip, let us know by leaving a reply in the comment section. Is there any question we can help you through? Feel free to sign up on our website to get the most professional advice from our experts.

Capacitor start motor

Working Principle

When the stator windings are energized from a 1-phase supply, the main winding and the starting winding, carry two different currents. There is a 90° time-phase difference and 90° space difference between the two currents. These two currents produce a rotating magnetic field that starts the motor.

The main and auxiliary windings are connected in parallel during motor starting. A start capacitor stays in the circuit long enough to rapidly bring the motor up to a predetermined speed, which is usually about 70 to 80 % of the full speed.



Then the auxiliary winding is disconnected from the supply often by a centrifugal switch and the motor remains powered by a single winding creating a pulsating magnetic field. In this sense, the auxiliary winding in this design can be regarded as a starting winding, since it is only used during motor startup.

Note that the motor will not work properly if the centrifugal switch is broken. If the switch is always open, the start capacitor is not part of the circuit, so the motor does not start. If the switch is always closed, the start capacitor is always in the circuit, so the motor windings will likely burn out. If a motor does not start, the capacitor is far more likely the problem than the switch.

Performance characteristics

The starting winding of a capacitor starts motor heats up less quickly and is well suited to applications involving either frequent or prolonged starting periods.

The power rating of such motors lies between 120 W and 7-5 kW.

The capacitor starts motor develops a much higher starting torque of about 3 to 4.5 times of the full load torque. To obtain a high starting torque, the two conditions are essential that are as follows:

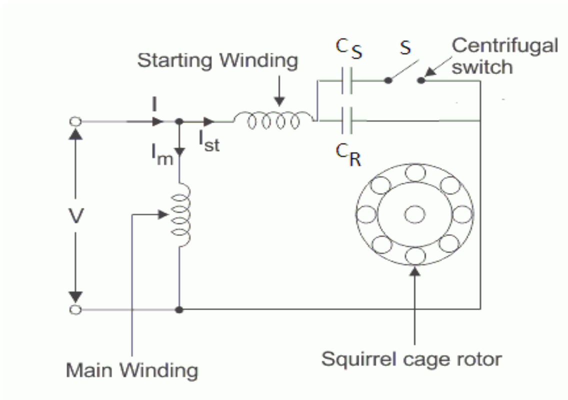
The Starting capacitor value must be large.

The valve of the starting winding resistance must be low.

Capacitor start , capacitor run motor

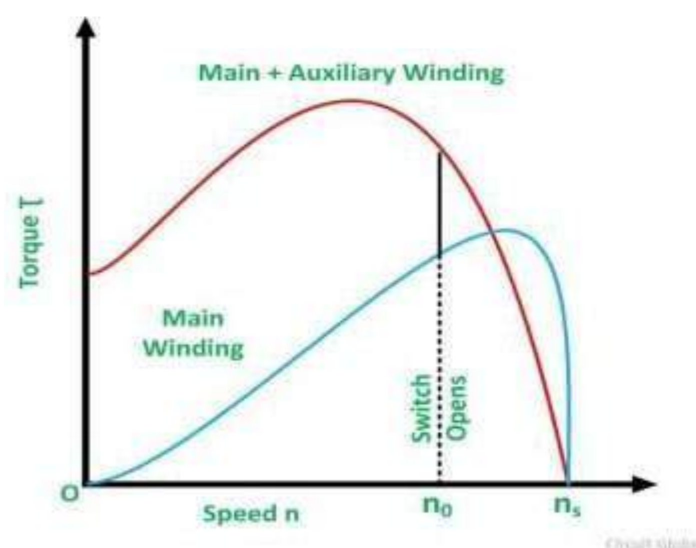
Working principle

Two capacitors are used in capacitor start capacitor run motor or two value capacitor motor, one for starting purpose and other for running purpose. Starting purpose capacitor is of electrolytic type and is disconnected from the supply when the motor attains 75% of synchronous speed with the help of centrifugal switch S, connected in series with Cs. The value of the two capacitors is different. Starting capacitor Cs, which is



electrolytic type, is of high value.

Characteristics of Capacitor Start Capacitor Run Motor



The capacitor start capacitor run motor gives the best running as well as starting conditions. Such motors operate as two-phase motors giving the best performance.

Starting torque is high, starting current is reduced and gives better efficiency, better p.f. The only disadvantage is high cost.

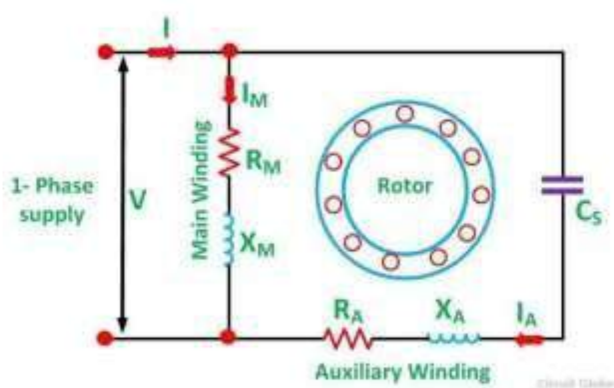
Direction can be reversed by interchanging the connection of supply to either of the main winding or starting winding.

Permanent capacitor type motor

Working principle

To use a single-phase power supply available in residential homes to drive a motor, there is a need for a mechanism to start the motor rotating. A PSC motor does this by having separate main and secondary windings (as shown in the diagram), with the main winding connected directly to the power supply and the secondary windings connected via a capacitor.

When the power supply is turned on, the current flows first in the main winding and then, with a short delay due to the capacitor, in the secondary winding. This difference in the main and secondary winding currents takes the form of a phase difference (meaning their waveforms are offset from one another on the time axis), causing the peak magnetic field to alternate between the two windings and thereby generating a torque that starts the motor rotation. When the stator windings are energised from a 1-phase supply, the main winding carries current Im and the starting winding carries current Ia. The capacitor Ca is connected in parallel during motor starting and running when the motor reaches full speed.



Applications

The various applications of the split motor are as follows:

Used in fans and blowers in heaters and air conditioners.

Used in refrigerator compressors.

Used in office machinery

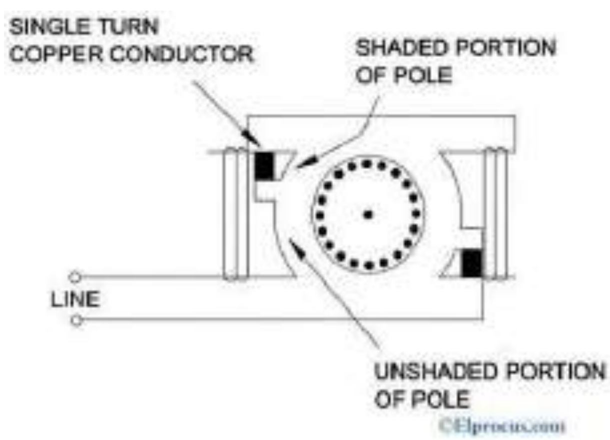
Shaded pole motor

Working principle

When power is supplied to the stator, flux is induced in the main part of the pole. This flux induces a voltage in the shading coil. This acts as a secondary winding. By Lenz's law, the current direction should be in such a way that, is opposing the flux entering into the coil. This acts as a secondary winding of a transformer.

Working of Shaded Pole Motor

In the core, when a single phase is applied an alternating flux is generated. This flux links with the shaded coil in fraction amounts. Then voltage gets induced in the coil due to the variation in the flux linking. Hence, the shaded portion is short-circuited due to which it produces the circulating current in it. In such a way, the direction is opposing the main flux.



The main core flux is opposed by the flux in the ring that is developed by the circulating current. Hence, flux is induced in the shaded portion of the motor along with the unshaded portion with a phase difference, which is lagging behind the unshaded pole flux. There is also a space displacement that is less than 90 degrees between a shaded ring flux and the main motor flux. Due to this space displacement, a rotating magnetic field is produced which leads to a torque on the cage motor. In order to obtain reversal in the direction of rotation, we have to provide two shading coils.

Applications

The applications of the shaded pole motor include the following.

Relays, Fans and other small devices due to its low cost

Exhaust fans

Hairdryers

Table fans

Cooling fans

Refrigerators

Air conditioners

Projectors

Record players

Tape recorders

Photocopying machines and many more.

5	Questions	1. State the applications of shaded pole motor. 2. Write the performance characteristics of different types of 1-φ induction motor.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-44

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	1-φ Induction motor , Method of changing the direction of rotation of 1-φ induction motor
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: Describe method changing the direction of rotation
3	Tools used	Chalk, Blackboard
4	Lecture Description	**Method of changing the direction of rotation of 1-φ induction motors** To reverse rotation on a single phase capacitor start motor, you will need to reverse the polarity of the starter winding. This will cause the magnetic field to change directions, and the motor will follow. In order to achieve this, you can swap the connections on either end of the winding. Always reverse the wires that lead to the starter winding. Note that most motors, if still labeled with manufacturer instructions, will indicate that they are non-reversible. If this is the case, it's likely because the wires you would need to access are housed within the motor. If your unit has this warning, it's easiest not to proceed. If, however, your motor is reversible, you might notice that it provides instructions for reversal. Usually, these instructions will tell you which wires to change. For instance, some units might have you switch the red and green leads in the terminal box at the end of the casing. In other instances, the blue and yellow wires may need to be swapped. The color of the wires is highly individual and depends upon the brand and model of motor you have. It's best to read the instructions provided by the manufacturer before proceeding. If you have determined which wires can be reversed, use a flathead screwdriver to remove the motor's end caps. Your motor may have a junction box instead. In either case, you'll need to gain access to the terminals where the indicated wires

		are attached. You may be able to use needle-nosed pliers to remove the wires and switch the terminals to which they attach. Some systems use nuts to connect the wires. If that's the case for you, you'll need to use a nut driver to gain access to the terminals. Once you've finished switching the wires, replace the end cap or close the junction box to the motor. Restore power to the circuit and test the motor to ensure the magnetic field has switched and that it is indeed now spinning in the opposite direction.
5	Questions	Explain the method of changing the direction of rotation of 1-φ induction motor.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-45

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	Commutator motors , Construction , Working principle , running characteristics , & applications of 1-φ series motor.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Define commutator 2. Say Constructional details of 1-φ series motor.
3	Tools used	Chalk, Blackboard
4	Lecture Description	**Commutator** A commutator is a rotary electrical switch in certain types of electric motors and electrical generators that periodically reverses the current direction between the rotor and the external circuit. It consists of a cylinder composed of multiple metal contact segments on the rotating armature of the machine. Two or more electrical contacts called "brushes" made of a soft conductive material like carbon press against the commutator, making sliding contact with successive segments of the commutator as it rotates. The windings (coils of wire) on the armature are connected to the commutator segments. **Single phase series motors**

The single-phase series motor is a commutator-type motor. If the polarity of the line terminals of a dc series motor is reversed, the motor will continue to run in the same direction. Thus, it might be expected that a dc series motor would operate on alternating current also. The direction of current through the armature[Taϕla]. direction of the torque developed in a dc series motor is determined by both filed polarity.

Construction

The construction of the ac series motor is very identical to the dc series motor except for some changes as given follows:

The magnetic circuit must be laminated.

Field circuit must be designed for much lower reactance than dc motor field in order to reduce the reactance voltage drop of field and improve power factor.

Distributed compensation winding is required to minimize the reactance of armature winding.

The compensating winding may be connected in series with the field and armature winding called conductively compensated and conductively compensation required armature which is designed to operate on ac and dc voltage.

If compensation winding short-circuited to itself and received excitation voltage by transformer action called inductively coupled

Operation

Let a dc series motor be connected across a single-phase ac supply. Since the same current flows through the field winding and the armature, it follows that ac reversals from positive to negative, or from negative to positive, will simultaneously affect both the field flux polarity and the current direction through the armature. This means that the direction of the developed torque will remain positive, and rotation will continue in the same direction. Thus, a series motor can run both on dc and ac.

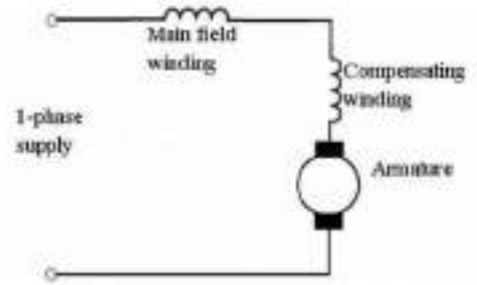
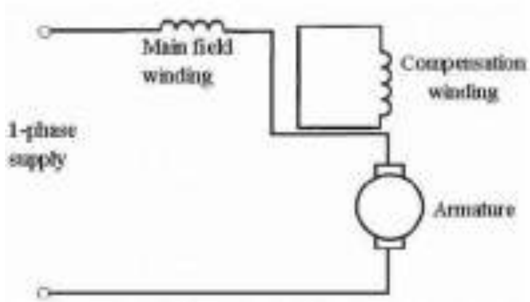


Fig 4.16



However, a series motor which is specifically designed for dc operation suffers from the following drawbacks when it is used on single-phase ac supply:

1. Its efficiency is low due to hysteresis and eddy-current losses.

2.The power factor is low due to the large reactance of the field and the armature winding.

3. The sparking at the brushes is excessive.

In order to overcome these difficulties, the following modifications are made in a D.C. series motor that is to operate satisfactorily on alternating current:

1. The field core is constructed of a material having low hysteresis loss. It is laminated to reduce eddy-current loss.

2. The field winding is provided with small number of turns. The field-pole areas is increased so that the flux density is reduced. This reduces the iron loss and the reactive voltage drop.

3.The number of armature conductors is increased in order to get the required torque with the low flux.

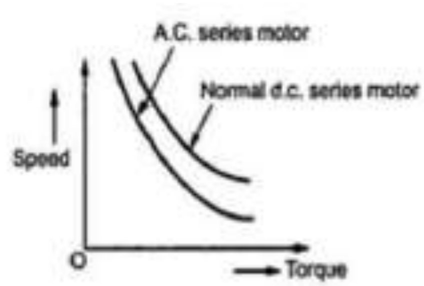
4.In order to reduce the effect of armature reaction, thereby improving commutation and reducing armature reactance, a compensating winding is used.

The compensating winding is put in the stator slots. The axis of the compensating winding is 90 (electrical) with the main field axis. It may be connected in series with both the armature and field as shown in Fig: In such a case the motor is conductively compensated.

The compensating winding may be short circuited on itself, in which case the motor is said to be inductively compensated shown in Fig.

Characteristics

The characteristics of single-phase series motor are very much similar to those of D.C. series motors, but the series motor develops less torque when operating from an a.c. supply than when working from an equivalent D.C. supply [Fig:]. The direction of rotation can be changed by interchanging connections to the field with respect to the armature as in D.C. series motor.



Applications

There are numerous applications where single-phase ac series motors are used, such as hair dryers, grinders, table-fans, blowers, polishers, kitchen appliances etc. They are also used for many other purposes where speed control and high values of speed are necessary.

5	Questions	1. Explain construction and operation of single phase series motor. 2. Explain characteristics and applications of single phase series motors.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-46

<table>
<tr><td colspan="3">Subject -Energy Conversion -||
Semeste-5th</td></tr>
<tr><td>1</td><td>Topic</td><td>Commutator motors , Construction, Working principle & applications of universal motors</td></tr>
<tr><td>2</td><td>Intended Learning Outcome</td><td>On successful completion of the Lecture students are able to:
1. Define universal motors.
2. Explain its construction.</td></tr>
<tr><td>3</td><td>Tools used</td><td>Chalk, Blackboard & Projector</td></tr>
<tr><td>4</td><td>Lecture Description</td><td>

Universal motors

The universal motor is a type of electric motor that can operate on either AC or DC power and uses an electromagnet as its stator to create its magnetic field. It is a commutated series-wound motor where the stator's field coils are connected in series with the rotor windings through a commutator.

Construction

Generally , universal motors are manufactured in two types :

1. Concentrated pole , non compensated type (low power rating)

2. Distributed field compensated type (high power rating)

The non compensated motor has two salient poles and is just like a 2 pole series dc motor except that whole of its magnetic path is laminated. The laminated stator is necessary because the flux is alternating when motor is operated from ac supply. The armature is of wound type and similar to that of a laminated core having either straight or skewed slots and a commutator to which the leads of the armature winding are connected. The distribution field compensated type motor has a stator core similar to that of a split phase motor and a wound armature similar to that of a small dc motor. The compensating winding is used to reduce the reactance voltage present in the armature when motor runs on ac supply. This voltage is caused by the alternating flux by transformer action.

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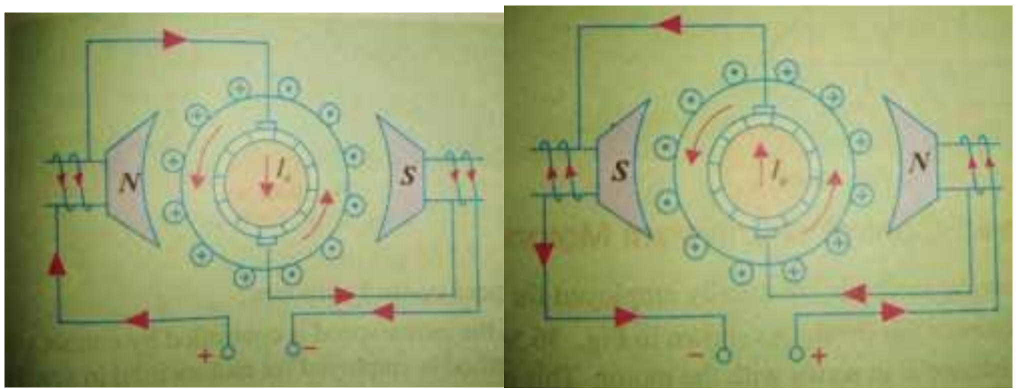
In a 2-pole non compensated motor , the voltage induced by transformer action in a coil during its communication period is not sufficient to cause any serious communication trouble. Moreover , high resistance brushes are used to aid commutation.

Working



(a). (b)

Some motors develop unidirectional torque , regardless of whether they operate on dc or ac supply. The production of unidirectional torque ,when the motor runs on ac supply can be easily understood from above fig.(a). The motor works on the same principle as a dc motor i.e. force between the main pole flux and the current carrying armature conductors. This is true regardless of whether the current is alternating or direct as shown in fig.(b).

Applications

Universal motors are used in vacuum cleaners where actual motor speed is the load speed. Other applications where motor speed is reduced by a gear train are : drink and food mixers , portable drills and domestic sewing machines etc.

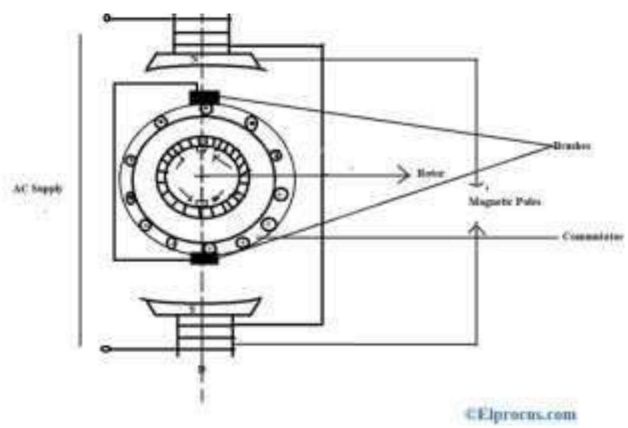
5	Questions	1. Explain principle of operation of universal motors. 2. State the applications of universal motors.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-48

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	Commutator motors , Working of repulsion start motor , Repulsion start induction run motor & Repulsion induction motor.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: Explain working principle of different types of repulsion motors.
3	Tools used	Chalk, Blackboard & Projector
4	Lecture Description	

Repulsion start motor

A repulsion motor is a single-phase electric motor that operates by providing input AC. The main application of repulsion motor is electric trains. It starts as a repulsion motor and runs as an induction motor, where the starting torque should be high for a repulsion motor and very good running characteristics for an induction motor. In repulsion motor direction of rotation of the motor is the same as that of brush shift.



Working principle

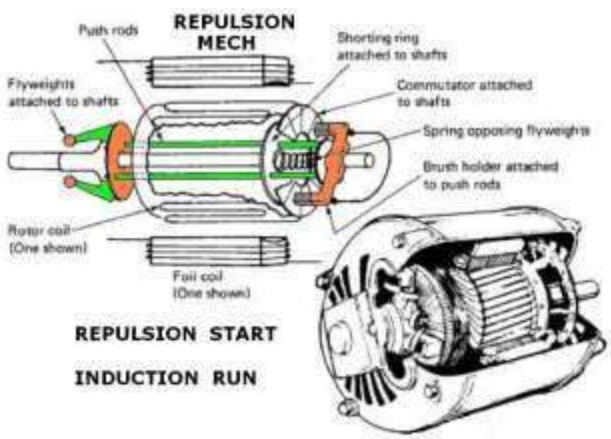
The basic principle behind the working of repulsion motor is that “similar poles repel each other.” This means two North poles will repel each other. Similarly, two South poles will repel each other. When the repulsion motor winding is supplied with single-phase AC, it produces a magnetic flux along the direct axis. WhenT this magnetic flux links with the rotor winding, it creates an EMF. Due to this EMF, a rotor current is produced. This rotor current in turn produces a magnetic flux that is directed along the brush axis due to commutator assembly. Due to the interaction of stator and rotor produced fluxes, an electromagnetic torque is produced.

Repulsion start induction run motor

The repulsion-start induction-run motor, starts as a repulsion motor and runs as an Induction motor. The repulsion start motor rotor is wound similar to a direct current armature. The stator is energized creating an alternating magnetic field that runs through the rotor and induces a current in the rotor windings.

Working principle

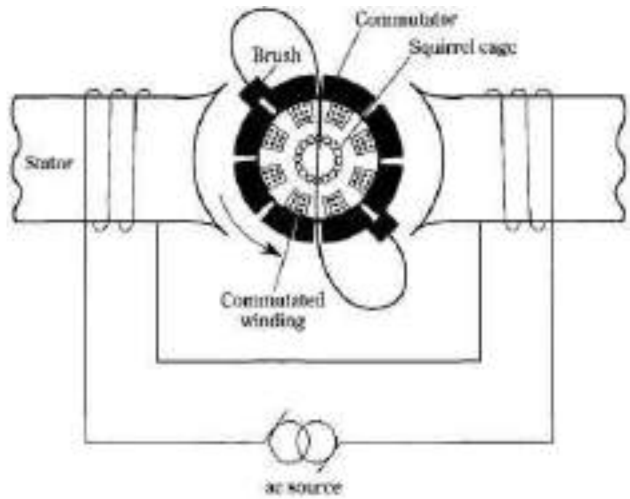
The basic principle behind the working of repulsion motor is that “similar poles repel each other.” This means two North poles will repel each other. Similarly, two South poles will repel each other. When the repulsion motor winding is supplied with single-phase AC, it produces a magnetic flux along the direct axis.



Repulsion induction motor

In the repulsion induction motor, a high starting torque is due to repulsion motor action. When it is running, the repulsion-induction motor functions through a combination of repulsion motor and induction motor action.

Working

		 When a 1-phase AC supply is fed to the stator winding of the repulsion induction motor, the repulsion winding is activated since at starting the impedance of the repulsion winding is less than that of the squirrel cage winding. As a result, the motor is started as a repulsion motor with a corresponding high starting torque. As the speed of the motor increases, the current shifts from repulsion winding to the squirrel cage winding due to the decreasing impedance of the squirrel cage winding with increasing speed. Therefore, under the running conditions, the squirrel cage winding carries the greater part of the rotor current. Consequently, the repulsion motor action is shifted to the induction motor action and it is achieved without any switching arrangement. Form the above discussion, it is clear that the motor starts as a repulsion motor and when running, it functions through a combination of principle of induction (predominant) and repulsion.
5	Questions	1. Explain working of repulsion start motor. 2. What do you mean by repulsion start induction run motor. 3. State the principle of repulsion induction motor.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-49

Subject -Energy Conversion -|| Semester- 5th

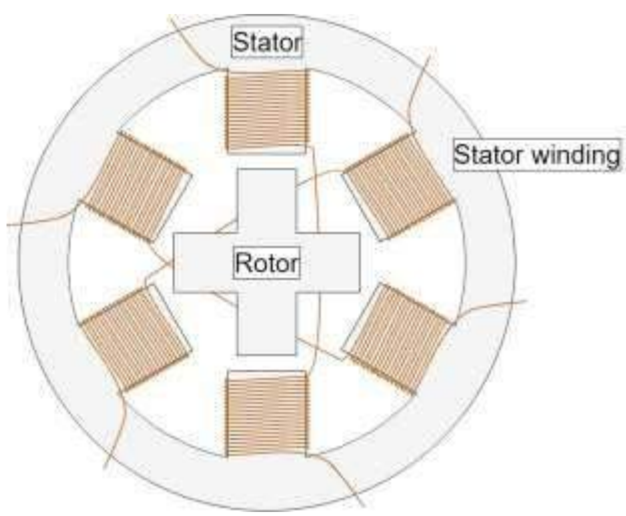
<table>
<tr><td>1</td><td>Topic</td><td>Special Electrical Machine , Principle of stepper motor</td></tr>
<tr><td>2</td><td>Intended Learning Outcome</td><td>On successful completion of the Lecture students are able to:
1. Define stepper motor.
2. State it's Principle.</td></tr>
<tr><td>3</td><td>Tools used</td><td>Chalk, Blackboard</td></tr>
<tr><td>4</td><td>Lecture Description</td><td>

Stepper motor

A stepper motor is an electromechanical device it converts electrical power into mechanical power. Also, it is a brushless, synchronous electric motor that can divide a full rotation into an expansive number of steps. The motor's position can be controlled accurately without any feedback mechanism, as long as the motor is carefully sized to the application. Stepper motors are similar to switched reluctance motors. The stepper motor uses the theory of operation for magnets to make the motor shaft turn a precise distance when a pulse of electricity is provided. The stator has eight poles, and the rotor has six poles. The rotor will require 24 pulses of electricity to move the 24 steps to make one complete revolution. Another way to say this is that the rotor will move precisely 15° for each pulse of electricity that the motor receives.

Principle of stepper motor

Figure shows a drawing representing the section of the motor is shown, where the rotor is a variable-reluctance iron core.



Cross-Section of a Stepper Motor

The basic working principle of the stepper motor is the following: By energizing one or more of the stator phases, a magnetic field is generated by the current flowing in the coil and the rotor aligns with this field. By supplying different phases in sequence, the rotor can be rotated by a specific amount to reach the desired final position. Figure 2 shows a representation of the working principle. At the beginning, coil A is energized and the rotor

</td></tr>
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		is aligned with the magnetic field it produces. When coil B is energized, the rotor rotates clockwise by 60° to align with the new magnetic field. The same happens when coil C is energized. In the pictures, the colors of the stator teeth indicate the direction of the magnetic field generated by the stator winding.
5	Questions	Explain working principle of stepper motor with diagram.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-50

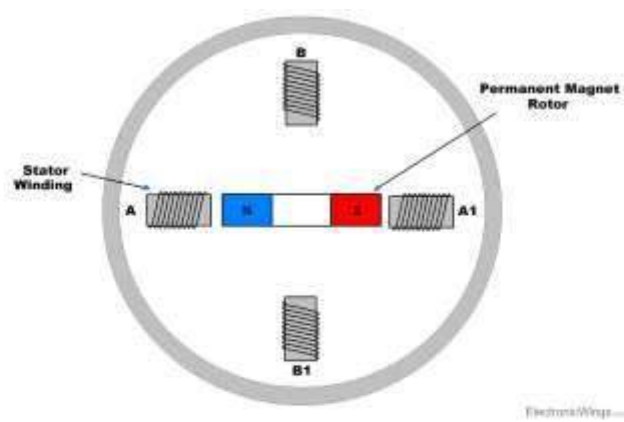
Subject -Energy Conversion -|| Semester- 5th		
1	Topic	Special Electrical Machine , Classification of stepper motor
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Identify the type of stepper motor.
3	Tools used	Chalk, Blackboard
4	Lecture Description	

Types Of Stepper Motors

There are three main types of stepper motors:

Permanent Magnet Stepper:

PM steppers have rotors that are constructed with permanent magnets, which interact with the electromagnets of the stator to create rotation and torque. PM steppers usually have comparatively low power requirements and can produce more torque per unit of input power.



Variable Reluctance Stepper:

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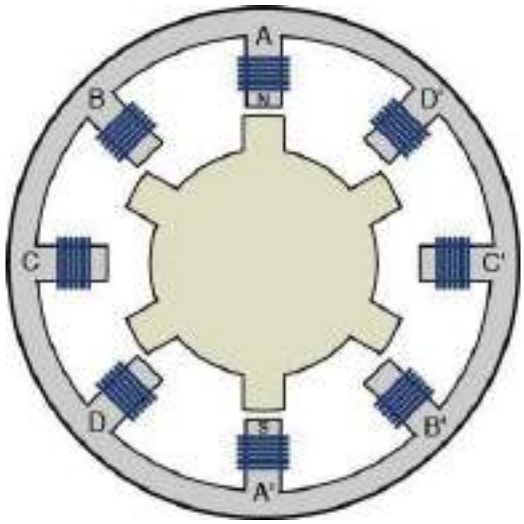
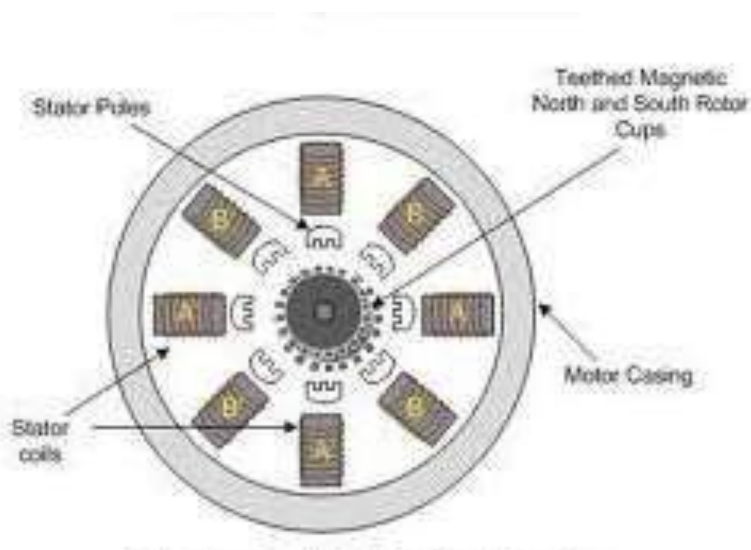
VR stepper rotors are not built with permanent magnets. Rather, they are constructed with plain iron and resemble a gear, with protrusions or “teeth” around the circumference of the rotor. The teeth lead to VR steppers that have a very high degree of angular resolution; however, this accuracy usually comes at the expense of torque.



Hybrid Syncronous Stepper:

HS stepper rotors use the best features of both PM and VR steppers. The rotor in an HS motor has a permanent magnet core, while the circumference is built from plain iron and has teeth. A hybrid synchronous motor, therefore, has both high angular resolution and high torque.



Notice the separate North and South Rotor Cup

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<tr><td>5</td><td>Questions</td><td>There are how many types of stepper motors. Explain about each.</td></tr>
<tr><td>6</td><td>Reference</td><td>https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html</td></tr>
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Lecture Plan-51

<table>
<tr><td colspan="3">Subject -Energy Conversion -||
Semester- 5th</td></tr>
<tr><td>1</td><td>Topic</td><td>Special Electrical Machine , Principle of variable reluctant stepper motor , permanent magnet stepper motor , & hybrid stepper motor.</td></tr>
<tr><td>2</td><td>Intended Learning Outcome</td><td>On successful completion of the Lecture students are able to:
State principle of different types of stepper motor.</td></tr>
<tr><td>3</td><td>Tools used</td><td>Chalk, Blackboard</td></tr>
<tr><td>4</td><td>Lecture Description</td><td>
<h1>Principle of VR Stepper Motor</h1>
The fundamental variable reluctance stepper motor working principle is that it is dependent on multiple reluctance locations of the rotor device. When the stator phases receive the voltage signal and get excites, there will be the creation of a magnetic field whose axis lines are across the poles.
And now, when the rotor attempts to revolve in such a path that it gains low reluctance. This revolution corresponds that a position magnetic field axis created by the stator is the same as the axis that passes across the rotor poles (any two of the poles).
<h1>Working Scenario</h1>
The variable reluctance stepper motor working can be easily explained by considering three cases. Let us know the working of this device in detail. Consider the below figure.
The working is explained as the three windings X, Y, and Z are connected in a series way and they are energized one after another using the three switches S1, S2, and S3.
<h1>Scenario 1</h1>
When the power supply is provided across the edges XX1, by closing the S1 switch. As there are magnetic poles in between the XX1 windings, because of attractional force between the magnetic poles, the rotor attempts to accomplish a low value of reluctance position. So, the 1 and 3 rotor axis attempts to be in alignment with the XX1 poles axis.
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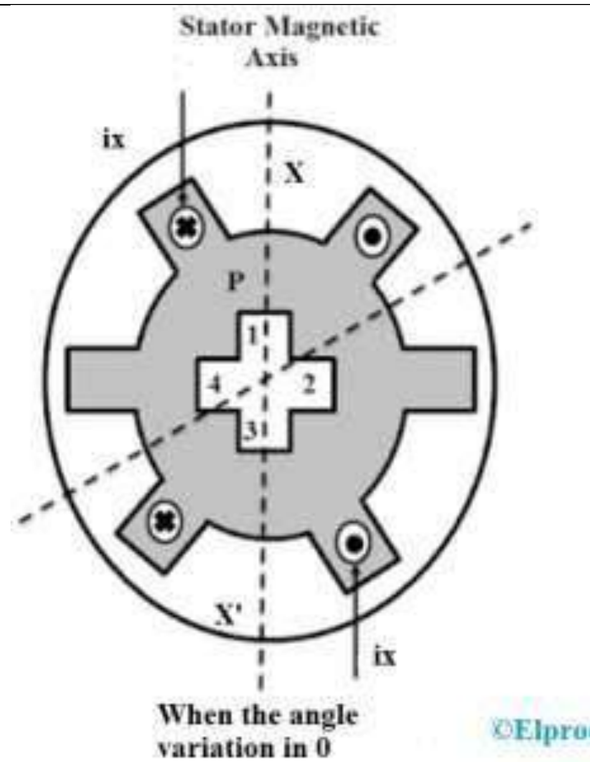
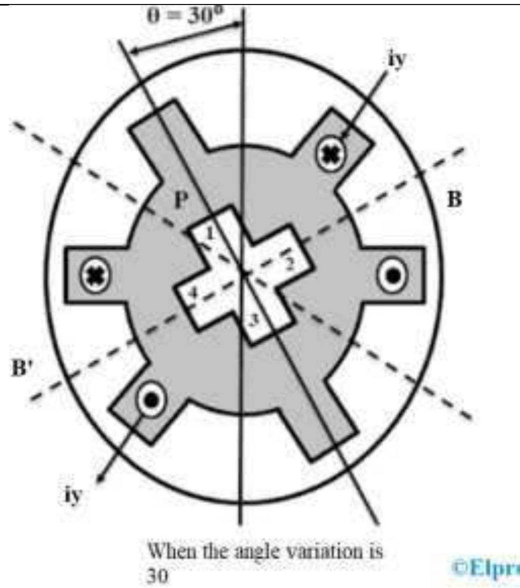


Working Scenario 1

Scenario 2

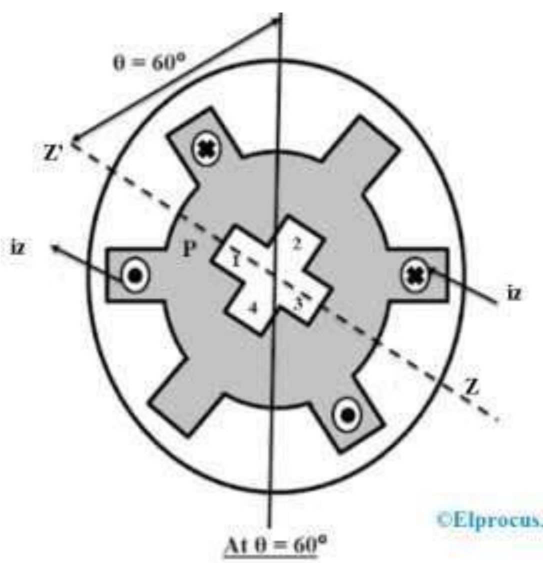
When the power supply is provided across the edges YY1, then there will be a modification in the stator poles magnetic axis. Now, the rotor attempts to accomplish a low reluctance direction in so creating the rotor movement. Here, 2 and 4 axis of the rotor poles gets so close to the YY1 windings. This creates rotor rotation and 2 and 4 rotor axis attempts to be in alignment with the YY1 poles axis. Hence, the rotor movement will move by 30 degrees.



Working Scenario 2

Scenario 3

In the same way, when the ZZ1 windings are energized by S3 making XX1 and YY as dis-connected. The magnetic poles of the rotor axis attempt to be alignment with the axis of the stator. Hence, the rotor movement will move by 30 degrees, so there will be a total rotation of 60 degrees from XX1 to ZZ1.



Working Scenario 3

With the successful execution of three phases in a corresponding manner, the motor finishes one revolution in 12 steps. And the rotor direction is based on the supply series provided to the stator phase. Then the torque generation that operates on the device has a direct proportion to the double of the phase current which is T α i2.

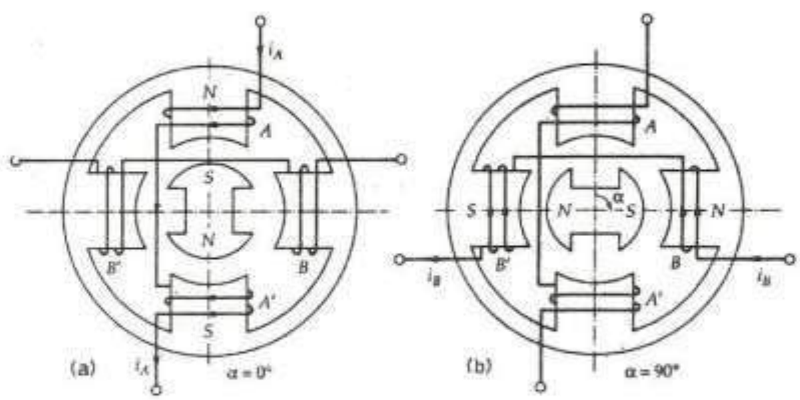
Principle of PM stepper motor

The working principle of the stepper motor is similar to that of a conventional motor. It works on the principle of Lorentz Force law. According to which, whenever a current-carrying conductor is placed in a magnetic field, it experiences a force, due to the interaction of fluxes.

The flux which interacts is stator magnetic flux and rotor magnetic flux. The stator magnetic flux is created due to external excitations and the rotor magnetic flux is created due to permanent magnets. It is also to be noted that, the direction of the motor is governed due to Fleming's left-hand rule.

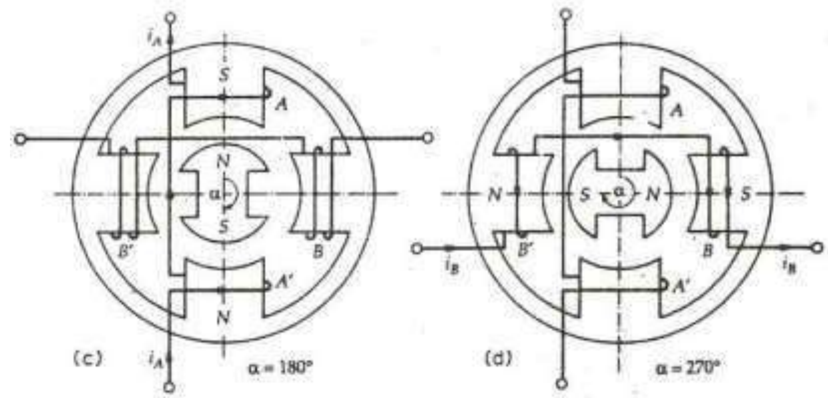
Working

The working permanent magnet stepper motor can be explained in the following modes



Mode 1– In this mode, the A phase of the stator poles are excited together with series winding to create two pairs of magnetic poles. It may be noted that, in this mode, the B phase is not excited at all. When the A phase is excited, it forms the North and South pole. At this moment, the rotor magnetic poles are attracted to the stator magnetic poles.

Mode 2 – In this mode, the B phase of the stator poles are excited together with series winding to create two pairs of magnetic poles. It may be noted that, in this mode, the A phase is not excited at all. When the B phase is excited, it forms the North and South pole. At this moment, the rotor magnetic poles are attracted to the stator magnetic poles. Which makes the rotor rotate in the clockwise direction from Mode 1.

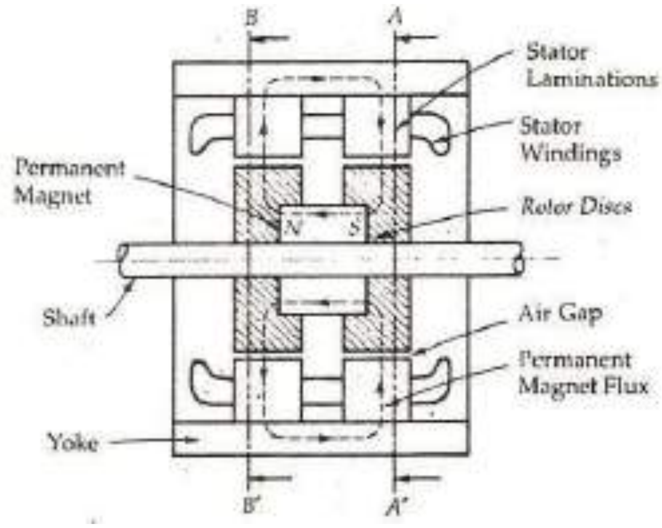


Mode 3 – Again In this mode, the A phase of the stator poles are excited together with series winding to create two pairs of magnetic poles. It may be noted that, in this mode, the B phase is not excited at all. When the A phase is excited, it forms the North and South pole. At this moment, the rotor magnetic poles are attracted to the stator magnetic poles. It makes the rotor rotate in the clockwise direction from mode 2.

Mode 4– Again In this mode, the B phase of the stator poles are excited together with series winding to create two pairs of magnetic poles. It may be noted that, in this mode, the A phase is not excited at all. When the B phase is excited, it forms the North and South pole. At this moment, the rotor magnetic poles are attracted to the stator magnetic poles. Which makes the rotor rotate in the clockwise direction from Mode 3.

In this manner, the rotor makes one complete revolution from mode 1 to mode 4.

Principle of Hybrid stepper motor

		 The working principle of the hybrid stepper motor is similar to the permanent magnet motor. In the above motor diagram, it has two phases, four poles, and six tooth rotor. Once the xx' is stimulated using DC supply then YY' cannot excite. So the poles of the rotor will be changed from one direction to another. Similarly, if the YY' phase is excited, then XX' will be turned off, so the position of poles will be changed. So the rotor in the motor will be changed into a new position in a counter-clockwise direction. If the YY' is excited oppositely then the higher pole changed to the south & lower pole changes to the north, after that the rotor will move in a clockwise direction. To operate the motor in the desired direction, a proper series of pulses need to given to the stator. So for each excitation, this will get protected into a new location. If the excitation is detached then this motor will maintain its locked state because of the excitation in a permanent magnet. The step- angle of this motor can be given as 30 degrees. Actually, the designing of these motors can be done using a number of rotor poles to attain high angular resolution.
5	Questions	Explain working of different types of stepper motor.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-52

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	Special Electrical Machine , Applications of stepper motor

2	Intended Learning Outcome	On successful completion of the Lecture students are able to: State the applications of stepper motor.
3	Tools used	Chalk, Blackboard
4	Lecture Description	**Applications of stepper motor** Such motors are used for operation control in computer peripherals , textile industry , IC fabrication , & robotics etc. Applications requiring incremental motion are typewriters , line printers , tape drives , floppy disk drives , numerically controlled machine tools , process control systems & X-Y plotters. Usually , position information , can be obtained simply by keeping count of the pulses sent to the motor thereby eliminating the need for expensive position sensors & feedback controls. Stepper motors also perform countless tasks outside the computer industry. It includes commerical , military & medical applications where these motors perform such functions as mixing , cutting , striking , metering blending & purging. They also take part in the manufacture of packed food stuffs , commerical end products & even the production of science fiction movies.
5	Questions	Write the uses of stepper motor.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-53

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	3-φ Transformer , Grouping of Winding & Advantages
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: State the function of 3-φ transformer
3	Tools used	Chalk, Blackboard and PPT

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<td>4</td>
<td>Lecture Description</td>
<td>

Grouping of Winding

The three phase transformer windings can be connected several ways. Based on the windings' connection, the vector group of the transformer is determined.
The Determination of vector group of transformers is very important before connecting two or more transformers in parallel. If two transformers of different vector groups are connected in parallel then phase difference exist between the secondary of the transformers and large circulating current flows between the two transformers which is very detrimental.
An ac voltage applied to a coil will induce a voltage in a second coil where the two are linked by a magnetic path. The relationship of the two voltages depends upon which ways round the coils are connected. The voltages will either be in-phase or displaced by 180 degree.
When 3 coils are used in a 3 phase transformer winding a number of options exist. The coil voltages can be in phase or displaced as above with the coils connected in star or delta and, in the case of a star winding, have the star point (neutral) brought out to an external terminal or not.
When Pair of Coil of Transformer have same direction than voltage induced in both coil are in same direction from one end to other end. When two coil have opposite winding direction than Voltage induced in both coil are in opposite direction.

Advantages of 3-φ transformer

Inexpensive:

Compared with single-phase transformers, 3-phase transformers are cheaper. It has a lower cost when compared to three units of a single-phase transformer.

Lightweight:

Three-phase transformers are light in weight and smaller in size than single-phase transformers, which means they take up less space.

Assembly in the least time:

Very easy to assemble. We can get a single-phase source from three-phase transformers, while the opposite is not valid.

Higher efficiency:

The three-phase transformer performs its functions more efficiently and delivers more power than a single-phase transformer.

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		Easier to install: Pre-wired 3-phase transformer is ready to install, making the installation very easy and smooth. **Easy transportation:** To provide the same output, the material used to fabricate the core of a 3-phase transformer is very little compared to that of 3 single-phase transformers. Therefore, they aren't only easy transportation but also lower shipping costs.
5	Questions	State the advantages of 3-φ transformer.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-54

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	3-φ transformer , Parallel operation of 3-φ transformer
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: Explain parallel operation of 3-φ transformer.
3	Tools used	Chalk, Blackboard
4	Lecture Description	**Parallel operation of 3-φ transformer** All the conditions which apply to the parallel operation of single phase transformers also apply to the parallel running of 3-phase transformers but with the following additions: 1. The voltage ratio must refer to the voltage of primary and secondary. It is obvious that this ratio may not be equal to the ratio of the number of turns per phase. For example , if V1 , V2 are the primary and secondary terminal voltages , then for Y/Δ connection , the turn ratio is $V2/(V1/\sqrt{3}) = \sqrt{3}V2/V1$.

		2. The phase displacement between primary and secondary voltages must be the same for all transformers which are to be connected for parallel operation. 3. The phase sequence must be the same. 4. All three transformers in the 3-phase transformer bank will be of the same construction either core or shell.
5	Questions	Explain parallel operation of 3-φ transformers.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-55

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	OMR-5
2	Intended Learning Outcome	On successful completion of the Lecture students are able to:
3	Tools used	Chalk, Blackboard
4	Lecture Description	
	Problems-	
5	Questions	
6	Reference	

Lecture Plan-56

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	3-φ transformers , Tap charger (on/off load tap charging)
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: Define tap charger.
3	Tools used	Chalk, Blackboard
4	Lecture Description	A tap changer is a mechanism in transformers which allows for variable turn ratios to be selected in distinct steps. This is done by connecting to a number of access points known as taps along either the primary or secondary winding.

Modern tap changer of a transformer.

Tap changers exist in two primary types,[1] no-load tap changers (NLTC), which must be de-energized before the turn ratio is adjusted, and on-load tap changers (OLTC), which may adjust their turn ratio during operation. The tap selection on any tap changer may be made via an automatic system, as is often the case for OLTC, or a manual tap changer, which is more common for NLTC. Automatic tap changers can be placed on a lower or higher voltage winding, but for high-power generation and transmission applications, automatic tap changers are often placed on the higher voltage (lower current) transformer winding for easy access and to minimize the current load during operation.[2]

Tap changing

No-load tap changer

No-load tap changer (NLTC), also known as Off-circuit tap changer (OCTC) or De-energized tap changer (DETC), is a tap changer utilized in situations in which a transformer's turn ratio does not require frequent changing and it is permissible to de-energize the transformer system. This type of transformer is frequently employed in low power, low voltage transformers in which the tap point often may take the form of a transformer connection terminal, requiring the input line to be disconnected by hand and connected to the new terminal. Alternatively, in some systems, the process of tap changing may be assisted by means of a rotary or slider switch.

No load tap changers are also employed in high voltage distribution-type transformers in which the system includes a no load tap changer on the primary winding to accommodate transmission system variations within a narrow band around the nominal rating. In such systems, the tap changer will often be set just once, at the time of installation, although it may be changed later to accommodate a long-term change in the system voltage profile.

On-load tap changer

On-load tap changer (OLTC), also known as On-circuit tap changer (OCTC), is a tap changer in applications where a supply interruption during a tap change is unacceptable, the transformer is often fitted with a more expensive and complex on load tap changing mechanism. On load tap changers may be generally classified as either mechanical, electronically assisted, or fully electronic.

These systems usually possess 33 taps (one at centre "Rated" tap and sixteen to increase and decrease the turn ratio) and allow for ±10% variation[3] (each step providing 0.625% variation) from the nominal transformer rating which, in turn, allows for stepped voltage regulation of the output.

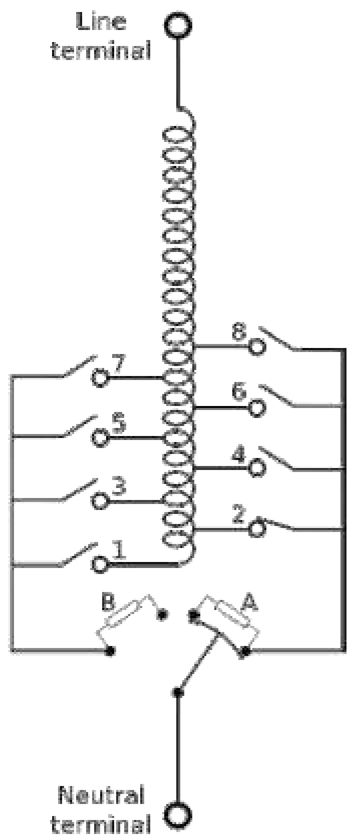
A mechanical On load tap changer (OLTC), also known as under-load tap changer (ULTC) design, changing back and forth between tap positions 2 and 3

Tap changers typically use numerous tap selector switches which may not be switched under load, broken into even and odd banks, and switch between the banks with a heavy-duty diverter switch which can switch between them under load. The result operates like a dual-clutch transmission, with the tap selector switches taking the place of the gearbox and the diverter switch taking the place of the clutch.

Mechanical tap changers

A mechanical tap changer physically makes the new connection before releasing the old using multiple tap selector switches but avoids creating high circulating currents by using a diverter switch to temporarily place a large diverter impedance in series with the short-circuited turns. This technique overcomes the problems with open or short circuit taps. In a resistance type tap changer, the changeover must be made rapidly to avoid overheating of the diverter. A reactance type tap changer uses a dedicated preventive autotransformer winding to function as the diverter impedance, and a reactance type tap changer is usually designed to sustain off-tap loading indefinitely.

In a typical diverter switch, powerful springs are tensioned by a low power motor (motor drive unit, MDU), and then rapidly released to effect the tap changing operation. To reduce arcing at the contacts, the tap changer operates in a chamber filled with insulating transformer oil, or inside a vessel filled with pressurized SF_6 gas. Reactance-type tap changers, when operating in oil, must allow for the additional inductive transients generated by the autotransformer and commonly include a vacuum bottle contact in parallel with the diverter switch. During a tap change operation, the potential rapidly increases between the two electrodes in the bottle, and some of the energy is dissipated in an arc discharge through the bottle instead of flashing across the diverter switch contacts.

		Some arcing is unavoidable, and both the tap changer oil and the switch contacts will slowly deteriorate with use. To prevent contamination of the tank oil and facilitate maintenance operations, the diverter switch usually operates in a separate compartment from the main transformer tank, and often the tap selector switches will be located in the compartment as well. All of the winding taps will then be routed into the tap changer compartment through a terminal array. One possible design (flag type) of on load mechanical tap changer is shown to the right. It commences operation at tap position 2, with load supplied directly via the right hand connection. Diverter resistor A is short-circuited; diverter B is unused. In moving to tap 3, the following sequence occurs:  Switch 3 closes, an off-load operation. Rotary switch turns, breaking one connection and supplying load current through diverter resistor A. Rotary switch continues to turn, connecting between contacts A and B. Load now supplied via diverter resistors A and B, winding turns bridged via A and B.

		Rotary switch continues to turn, breaking contact with diverter A. Load now supplied via diverter B alone, winding turns no longer bridged. Rotary switch continues to turn, shorting diverter B. Load now supplied directly via left hand connection. Diverter A is unused. Switch 2 opens, an off-load operation. The sequence is then carried out in reverse to return to tap position 2. **Solid-state tap changer** This is a relatively recent development which uses thyristors both to switch the transformer winding taps and to pass the load current in the steady state. The disadvantage is that all non-conducting thyristors connected to the unselected taps still dissipate power due to their leakage currents and they have limited short circuit tolerance. This power consumption can add up to a few kilowatts which appears as heat and causes a reduction in overall efficiency of the transformer; however, it results in a more compact design that reduces the size and weight of the tap changer device. Solid state tap changers are typically employed only on smaller power transformers.
5	Questions	1. Write the function of tap charger. 2. What do you mean by solid state tap charger.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-57

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	3-φ transformers , Maintenance scedule of power transformer
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: Say maintenance process of 3-φ transformer
3	Tools used	Chalk, Blackboard
4	Lecture Description	**What Is The Maintenance Of Transformers** **A p**ower transformer is the most costly and essential equipment piece of equipment within an electrical substation. As such it is desirable to perform various preventative maintenance activities to ensure the transformer maintains a high level of performance and a long functional life. A power transformer requires various routine maintenance tasks including measurement and testing of different parameters of the transformer. There are

two main types of maintenance of transformer. We perform one group on a routine basis (known as preventative maintenance), and the second group on an ad-hoc basis (i.e. as required).

That means for getting smooth performance from a transformer we have to perform some maintenance actions on a regular basis.

Transformer Maintenance

Some other types of maintenance for a transformer we perform only as they are required – known as emergency or breakdown transformer maintenance. But if one performs regular maintenance properly, this significantly reduces the chances of needing to perform such emergency maintenance. The regular checking and maintenance of transformer is also known as condition maintenance.

Hence by proper condition maintenance, one can avoid emergency and breakdown maintenance. That is why technical personnel should mainly concentrate on condition maintenance. As 100% of condition maintenance causes 0% breakdown of equipment.

There are many different preventative maintenance actions to be performed on a power transformer. They can be on a daily, monthly, yearly, quarterly, half-yearly, or yearly basis. Some transformer maintenance activities only need to be performed once in a 3 to 4 years interval.

Monthly Basis Maintenance of Transformer

Let us first discuss about the action to be taken on power transformer in monthly basis.

The oil level in oil cap under silica gel breather must be checked in a one-month interval. If it is found the transformer oil inside the cup comes below the specified level, oil to be top up as per specified level.

Breathing holes in silica gel breather should also be checked monthly and properly cleaned if required, for proper breathing action.

If the transformer has oil filled bushing the oil level of transformer oil inside the bushing must be vidually checked in the oil gage attached to those bushing. This action also to be done monthly basis.

If it is required, the oil to be filled in the bushing upto correct level. Oil filling to be done under shutdown condition.

Daily Basis Maintenance and Checking

There are three main things which to be checked on a power transformer on a daily basis:

Reading of MOG (Magnetic Oil Gauge) of main tank and conservator tank.

Color of silica gel in breather.

Leakage of oil from any point of a transformer.

In case of unsatisfactory oil level in the MOG, oil to be filled in transformer and also the transformer tank to be checked for oil leakage. If oil leakage is found take required action to plug the leakage. If silica gel becomes pinkish, it should be replaced.

Yearly Basis Transformer Maintenance Schedule

The auto, remote, manual function of cooling system that means, oil pumps, air fans, and other items engaged in cooling system of transformer, along with their control circuit to be checked in the interval of one year. In the case of trouble, investigate control circuit and physical condition of pumps and fans.

All the bushings of the transformer to be cleaned by soft cotton cloths yearly. During cleaning the bushing should be checked for cracking.

Oil condition of OLTC to be examined in every year. For that, oil sample to be taken from drain valve of divertor tank, and this collected oil sample to be tested for dielectric strength (BDV) and moisture content (PPM). If BDV is low and PPM for moisture is found high compared to recommended values, the oil inside the OLTC to be replaced or filtered.

All marshalling boxes to be cleaned from inside at least once in a year. All illumination, space heaters, to be checked whether they are functioning properly or not. If not, required maintenance action to be taken. All the terminal connections of control and relay wiring to be checked an tighten at least once in a year.

All the relays, alarms and control switches along with their circuit, in R&C panel (Relay and Control Panel) and RTCC (Remote Tap Changer Control Panel) to be cleaned by appropriate cleaning agent.

The pockets for OTI, WTI (Oil Temperature Indicator & Winding Temperature Indicator) on the transformer top cover to be checked and if required oil to be replenished.

		The proper function of Pressure Release Device and Buchholz relay must be checked annually. For that, trip contacts and alarm contacts of the said devices are shorted by a small piece of wire, and observe whether the concerned relays in remote panel are properly working or not. Insulation resistance and polarization index of transformer must be checked with battery operated megger of 5 KV range. Resistive value of earth connection and rizer must be measured annually with clamp on earth resistance meter. DGA or Dissolve Gas Analysis of transformer Oil should be performed, annually for 132 KV transformer, once in 2 years for the transformer below 132 KV transformer and in 2 years interval for the transformer above 132 KV transformer. The Action to be taken once in 2 years: The calibration of OTI and WTI must be carried once in two years. Tan & delta; measurement of bushings of transformer also to be done once in two years. **Maintenance of Transformer on Half Yearly Basis** The transformer oil must be checked half yearly basis that means once in 6 months, for dielectric strength, water content, acidity, sludge content, flash point, DDA, IFT, resistivity for transformer oil. In the case of a distribution transformer, as they are operating light load condition all the time of day remaining peak hours, so there are no maintenance required.
5	Questions	1. Write the maintenance schedule for 3-φ transformer. 2. Write the actions will be taken for maintenance once in 2 years.
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-58

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	
2	Intended Learning Outcome	On successful completion of the Lecture students are able to:

3	Tools used	Chalk, Blackboard
4	Lecture Description	
5	Questions	
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

Lecture Plan-59

Subject -Energy Conversion -|| Semester- 5th		
1	Topic	
2	Intended Learning Outcome	On successful completion of the Lecture students are able to:
3	Tools used	Chalk, Blackboard
4	Lecture Description	
5	Questions	
6	Reference	https://pdfcoffee.com/a-textbook-of-electrical-technology-sbltheraja-volume-iipdf-pdf-free.html

UNIT DISCUSSION- III, IV

1. What is V-curves of synchronous motor ?

2. What do you mean by infinite bus ?

3. Which type of alternator is used in hydroelectric power plants and why ?

4. What is the relation between electrical degrees and mechanical angle of an alternator ?

5. State the difference between coil pitch and pole pitch.

6. State pitch factor and distribution factor.

7. What is the function of damper bars ?

8. What is the principle of operation of synchronous motor ?

9. Why a starting winding is needed ?

10. In which rotor high starting torque is produced and why

11. What are the types of alternator ?

12. What frequency voltage must be applied to a 20-pole synchronous motor so that it runs at a speed of 300 rpm ?

13. Can a three phase IM never runs at synchronous speed ?

14. How frequency of rotor current is related to slip ?

15. Why are alternator rated in kVA , not in kW ?

16. Why single phase induction motor is not self starting ?

17. If a single phase induction motor fails to start but gives a humming sound , what is the reason behind it ?

18. What are the two advantages of short pitched coil in an alternator wdg.

19. Calculate the value of 'Kd' for a 36-slots , 4-poles single layer three phase alternator.

20. For traction purpose which motors are used ?

21. State two applications of DC series motor.

22. State the difference between coil pitch and pole pitch.

23. State pitch factor and distribution factor.

24. What is the function of damper bars ?

25. What is the principle of operation of synchronous motor ?

26. Why a starting winding is needed ?

27. In which rotor high starting torque is produced and why ?

28. What are the types of alternator ?

29. What frequency voltage must be applied

30. Why rotor slots of an induction motor are skewed ?

31. Write down the advantages of stationary armature of alternator ?

32. What do you mean by voltage regulation of an alternator ?

33. What is damper bar or damper winding ?

34 State the applications of synchronous motor.

35. How the direction of rotation of a single phase induction motor can be reversed ?

36. Why parallel operation of alternators needed ?

37. What is hunting ?

38. What is the function of damper bars ?

39. What is the principle of operation of synchronous motor ?

40. Why a starting winding is needed ?

41. In which rotor high starting torque is produced and why

42. What are the types of alternator ?

43. What frequency voltage must be applied to a 20-pole synchronous motor so that it runs at a speed of 300 rpm ?

44. Can a three phase IM never runs at synchronous speed ?

45. How frequency of rotor current is related to slip ?

45. Why are alternator rated in kVA , not in kW ?

46. What is V-curves of synchronous motor ?

47. What do you mean by infinite bus ?

48. Which type of alternator is used in hydroelectric power plants and why ?

49. What is the relation between electrical degrees and mechanical angle of an alternator ?

50. State the difference between coil pitch and pole pitch.

51. State pitch factor and distribution factor.

52. What is the function of damper bars ?

53. What is the principle of operation of synchronous motor ?

54. Why a starting winding is needed ?

55. What is the principle of operation of synchronous motor ?6. Why a starting winding is needed ?

56. In which rotor high starting torque is produced and why ?

57. What are the types of alternator ?

58. What frequency voltage must be applied

59. Why rotor slots of an induction motor are skewed ?

Lecture Plan

Subject - UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	ELECTROLYTIC PROCESS Definition and Basic principle of Electro Deposition.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Principle of electrodeposition 2. Faradays law of electrolysis 3. Metal extraction
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	**Definition and Basic principle of Electro Deposition.** Electro deposition is the process of coating a thin layer of one metal on top of different metal to modify its surface properties. It is done to achieve the desire electrical and corrosion resistance, reduce wear &friction, improve heat tolerance and for decoration. **Electroplating Basics** Figure- 1, schematically illustrates a simple electrochemical plating system. The —electro part of the system includes the voltage/current source and the electrodes, anode and cathode, immersed in the —chemical part of the system, the electrolyte or plating bath, with the circuit being completed by the flow of ions from the plating bath to the electrodes. The metal to be deposited may be the anode and be ionized and go into solution in the electrolyte, or come from the composition of the plating bath. Copper, tin, silver and nickel metal usually comes from anodes, while gold salts are usually added to the plating bath in a controlled process to maintain the composition of the bath. The plating bath generally contains other ions to facilitate current flow between the electrodes. The deposition of metal takes place at the cathode. The overall plating process occurs in the following sequence: 1. Power supply pumps electrons into the cathode. 2. An electron from the cathode transfers to a positively charged metal ion in the solution and the reduced metal plates onto the cathode. 3. Ionic conduction through the plating bath completes the circuit to the anode. 4. At the anode two different processes take place depending on whether the anode material is soluble, the source of the metal to be plated, or insoluble, inert. If the anode material is soluble, a metal atom gives up an electron and goes into the solution as a positively charged metal ion replenishing the metal content of the plating bath. If the anode is inert a negatively charged ion from the plating bath gives up an electron to the anode. 5. The electron flows from the anode to the power supply completing the circuit. The deposition of metal at the cathode requires an electron so the rate of deposition depends on the flow of electrons, that is, the current flowing from the rectifier. The thickness of the deposit, therefore, depends on the current and the length of time the current is applied. This relationship is a result of Faraday's law which relates the weight of a substance produced by an anode or cathode electrode reaction during electrolysis as being directly proportional to the

quantity of electricity passed through the cell.

Faraday's Laws of Electrolysis

From his experiments, Faraday deduced two fundamental laws which govern the phenomenon of electrolysis. These are:

i. First Law. The mass of ions liberated at an electrode is directly proportional to the quantity of electricity i.e. charge which passes through the electrolyte

Or

The weight of a substance liberated from an electrolyte in a given time is proportional to the quantity of electricity passing through the electrolyte.

That is W Q It , where I is the current and t is the time. W = Zit Where Z is a constant called electro-chemical equivalent. If I = 1 ampere and T= one second then, Z = W , which gives a definition of Z. The electro-chemical equivalent of a substance is the amount of that substance by weight liberated in unit time by unit current.

Second Law. The masses of ions of different substances liberated by the same quantity of electricity are proportional to their chemical equivalent weights.

or,

ii. If the same current flows through several electrolytes, the weights of ions liberated are proportional to their chemical equivalents. The chemical equivalent of a substance is the weight of the substance which can displace or combine with unit weight of hydrogen .Obviously, the chemical equivalent of hydrogen is 1 by definition.

DEFINITIONS

1. **Current Efficiency**

   On account of the impurities which cause secondary reactions, the quantity of a substance liberated is less than that calculated from faraday's Law. Current efficiency is the ratio of the actual mass of a substance liberated from an electrolyte by the passage of current to the theoretical mass liberated according to Faraday's law. Current efficiency can be used in measuring electro deposition thickness on materials in electrolysis. Current efficiency is also known as faradic efficiency, faradic yield and columbic efficiency.

2. **Energy Efficiency**

   On account of secondary reactions, the voltage actually required for the deposition or liberation of metal is higher than the theoretical value which increases the actual energy required.

   Energy efficiency is defined as $=\frac{Theoritical\ energy}{Actual\ energy\ required}$

   It is a process by which a metal is deposited over another metal or non-metal. Electro-plating is a very common example of such process.

   i. Conditions have to be provided so that the deposit will be fine grained and will have a smooth appearance. The factors which affect the electro-deposition of metals are :

(i) Current Density
(ii) Electrolyte concentration
(iii) Temperature
(iv) Addition agents
(v) Nature of electrolyte

(vi) Nature of the metal on which the deposit is to be made
(vii) Throwing power of the electrolyte

Current density

At low values of current density the ions are released at a slow rate and the rate of growth of nuclei is more than the rate at which the new nuclei form themselves. Electro-deposition depends upon the rate at which crystals grow and the rate at which fresh nuclei are formed. Therefore, at low current densities the deposit will be coarse and crystalline in nature. At higher values of current density the quality of deposit becomes more uniform and fine-grained on account of the greater rate of formation of nuclei. If the current density is so high that it exceeds the limiting value for the electrolyte hydrogen is released and spongy and porous deposit is obtained.

Electrolytic Concentration

This is more or less complementary to the first factor, i.e. current density, since by increasing the concentration of the electrolyte higher current density can be achieved. Increase of concentration tends to give better deposits and some people therefore favour it.

Temperature

The temperature of the electrolyte has two contradictory effects. One, at comparatively high temperature there is more diffusion and even at relatively high current density smooth deposits may be produced. Two, the rate of crystal growth increases the possibility of coarse deposits. At moderate temperatures the deposits are good. In chromium plating the temperature is maintained at C, and in nickel between 50^0C to 60^0C

Addition Agents

the quality of a deposit is improved by the presence of an addition agent which may be colloidal matter or an organic compound, otherwise the metal deposits in the form of large crystals and the surface becomes rough. Materials used as addition agents are gelatin, agar, glue, gums, rubber, alkaloids, sugar etc. The addition agents are supposed to be absorbed by crystal nuclei and prevent their growth into large crystals. The discharged ions start to build up new nuclei and the deposit of metal is fine-grained.

Nature of electrolyte

Smooth deposits are obtained from solutions having complex ions, e.g.,cyanides. Silver from nitrate solution forms a coarse deposit while from cyanide solution it forms a smooth deposit. Therefore, the formation of smooth deposit largely depends upon the nature of electrolyte used.

Nature of the metal on which deposit is to be made

This factor influences the growth of crystals since it is believed that the operation of crystals is in continuation of these in the base metal.

Throwing Power

The throwing power of an electrolyte may be regarded as the quality which produces a uniform deposit on a cathode having an irregular shape. Since the shape is irregular, The distance of the various parts of the cathode from the anode is not the same and therefore the conductance of the electrolyte is not the same for all parts of the cathode. The phenomenon of throwing power has not been clearly understood so far. In an electrolyte of low conductance, the current will concentrate on the parts of the cathode which are nearer the cathode resulting in poor throwing power. If the electrolyte has good conductance, the throwing power will also be good. One way to improve the

		throwing power is to keep a good distance between the cathode and the anode thereby providing more or less the same conductance for all parts of cathode. Presence of colloidal matter improves the throwing power but increase of temperature may produce the opposite effect. **Extraction of Metals** This is done in two ways: 1. The ore is treated with a strong acid to obtain a salt and the solution of such a salt is electrolyzed to liberate the metal. 2. When the ore in molten state is available it is electolysed in a furnace. **Extraction of Zinc** The ore consisting of zinc is treated with concentrated sulphuric acid, roasted and passed through other processes to get rid of impurities by precipitation. The zinc-sulphate solution is then electrolyzed. The cells consist of large lead-lined wooden boxes having aluminum cathodes and lead anodes. The current density is about 1000 amperes per square meter. Zinc is deposited on cathodes **Extraction of Aluminum** Ores of aluminum are bauxite cryolite. Bauxite is treated chemically and reduced to aluminum oxide and then dissolved in fused cryolite and electrolysed.The furnace is lined with carbon. The temperature of the furnace is about 1000^{o}C to keep the electrolyte in a fused state. Aluminum deposits at the cathode **Refining of Metals** Electrolytic extraction gives about 98 to 99 percent pure metal. Further refining is done by electrolysis. The anodes are made of the impure metal extracted from its ores and the electrolyte is a solution of the salt of the metal. Pure metal is deposited on the cathode.
5	Questions	1. What is electrolysis process? 2. What is faradays law of electrolysis? 3. What is metal extraction?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

Lecture Plan-2

Subject - UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	Important terms regarding electrolysis
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. What is electrolytic cells 2. What is electrochemical series
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	Electrolysis is a chemical process in which the direct electric current (I) is passed through an electrolyte containing ionic compounds, which can be in either the molten state or aqueous phase. During electrolysis, the anion moves toward the anode and the cation moves toward the cathode. The experimental setup in which electrolysis takes place is called an electrolytic cell. Electrolytic cells: The electrolysis reaction is occurring in an electrolytic cell. The reactions occurring in the

		electrolytic cell is always non-spontaneous. External electronic energy is used to drive this non-spontaneous reaction. In this cell, the anode is located at the right side and the cathode at the left side. At the anode, oxidation reaction occurs while at the cathode, reduction reaction occurs. Anions migrate towards the anode and get liberated. Likewise, cations migrate towards cathode and get deposited. If more than one cation or anion is present in the solution: The reduction half-reaction with the more positive reduction potential will occur at the cathode. The oxidation half-reaction with the more negative (less positive) reduction potential will occur at the anode. Electrochemical series: Electrochemical series is the series of arrangement of metals in the reduction potential of the metals. The table of standard electrode potential gives the reduction potential values for various half-reactions of various ions in the aqueous, gaseous, and solid states. The half-reaction of hydrogen has the reduction potential value of zero. It is used as the reference value for determining the reduction potential of the other half-reactions. In the table, the half-reactions with the negative values indicate the stronger reducing agent and weaker oxidizing agents, and the half-reactions with positive value indicate the stronger oxidizing agent and weaker reducing agent.
5	Questions	1. Explain electrolytic cells?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

Lecture Plan-3

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	Faradays Laws of Electrolysis.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. What is faradays Laws
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	Faraday's Laws of Electrolysis From his experiments, Faraday deduced two fundamental laws which govern the phenomenon of electrolysis. These are: First Law. The mass of ions liberated at an electrode is directly proportional to the quantity of electricity i.e. charge which passes through the electrolyte Or The weight of a substance liberated from an electrolyte in a given time is proportional to the quantity of electricity passing through the electrolyte. That is W Q It , where I is the current and t is the time. W = Zit Where Z is a constant called electro-chemical equivalent. If I = 1 ampere and T= one second

		then, Z = W , which gives a definition of Z. The electro-chemical equivalent of a substance is the amount of that substance by weight liberated in unit time by unit current. Second Law. The masses of ions of different substances liberated by the same quantity of electricity are proportional to their chemical equivalent weights. or, If the same current flows through several electrolytes, the weights of ions liberated are proportional to their chemical equivalents. The chemical equivalent of a substance is the weight of the substance which can displace or combine with unit weight of hydrogen .Obviously, the chemical equivalent of hydrogen is 1 by definition.
5	Questions	1. What is faradays 1st law of electrolysis? 2. What is faradays 2nd law of electrolysis?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg

Lecture Plan-4

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	Definitions of current efficiency, Energy efficiency
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. What is current efficiency 2. What is energy efficiency
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	**DEFINITIONS** **Current Efficiency** On account of the impurities which cause secondary reactions, the quantity of a substance liberated is less than that calculated from faraday's Law. Current efficiency is the ratio of the actual mass of a substance liberated from an electrolyte by the passage of current to the theoretical mass liberated according to Faraday's law. Current efficiency can be used in measuring electro deposition thickness on materials in electrolysis. Current efficiency is also known as faradic efficiency, faradic yield and columbic efficiency. **Energy Efficiency** On account of secondary reactions, the voltage actually required for the deposition or liberation of metal is higher than the theoretical value which increases the actual energy required. Energy efficiency is defined as = It is a process by which a metal is deposited over another metal or non-metal. Electro-plating is a very common example of such process. Conditions have to be provided so that the deposit will be fine grained and will

		have a smooth appearance. The factors which affect the electro-deposition of metals are : (i) Current Density (ii) Electrolyte concentration (iii) Temperature (iv) Addition agents (v) Nature of electrolyte (vi) Nature of the metal on which the deposit is to be made (vii) Throwing power of the electrolyte
5	Questions	1. Explain current efficiency? 2. Explain energy efficiency? 3. Which factors affect the Electro-deposition metal?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

Lecture Plan-5

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	Principle of Electro Deposition
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. State the function of moving target indicator. 2. Draw the block diagram of moving target indicator Radar. 3. Describe the working principle of moving target indicator radar.
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	Electrodeposition is a flexible low-cost method of fabrication of a wide variety of two- and three-dimensional materials such as coatings and films. The principles of the electrodeposition process are based on principles of electrochemical phenomena associated with the reduction or deposition of electroactive and accompanying species on the cathode surface. This would make the electrodeposition process more controllable if one considers the electrochemical principles into account for target purposes and applications. However, many empirical factors in designing and controlling of electrodeposition process exist. Anticorrosion coatings have been a major part of corrosion protection tools employed for years. Electrodeposition is one successful method to fabricate coatings. Thanks to remarkable advances in nanoscience and nanotechnology and exotic effects of the nanoscale size on properties and functionalities of materials,

		electrodeposition of nanocoatings has become an interesting subject of research recently. In this chapter, the electrodeposition of anticorrosion nanocoatings is described. After a brief overview of the electrodeposition, two types of nanocoatings including nanocomposites and nanocrystalline, their electrodeposition principles, and corrosion behavior based on conventional electrochemical studies are explained.
5	Questions	1. What is Clutter? 2. State the function of moving target indicator. 3. What is the meaning of stationary targets? 4. How many oscillators are used in MTI Radar and state the function of all? 5. State the function of mixer present in transmitter section.
6	Reference	

Lecture Plan-6

Subject -Advanced Communication Engineering Semester- 6th		
1	Topic	Factors affecting the amount of Electro Deposition.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. What is current 2. What is efficiency 3. What is strength of solution
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	Factors Effecting the Amount of Electro-Deposition 1. Time. Time is directly proportional to the quantity of electrodeposition. So, we can say that more mass will be deposited in more time and less mass will be deposited in less time conditions remain same. If the other conditions remain same. 2. Efficiency.

		Greater is the efficiency, greater is the quantity of metal deposited for a given time. 3. Current. The value of current is also directly proportional to the mass of metal deposited. Greater is the current, greater is the quantity of metal deposited while other conditions remaining the same. But, if we increase the current beyond a certain limit which is fixed for different metals separately, the metal deposited will be of different colour such as blackish, which is known as burnt metal. 4. Strength of solution. It the strength of solution is more, then the mass of metal deposited will be more as compared to the dilute solution of electrolyte if the other conditions remain the same.
5	Questions	1. . What is efficiency? 2. What is time ? 3. Explain strength of solution ?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

Lecture Plan-7

Subject -Advanced Communication Engineering Semester- 6th		
1	Topic	Factors governing the electro deposition
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. What is nature of electrolyte 2. Explain current density 3. What is temperature
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	Factor # 1. Nature of Electrolyte: The formation of smooth deposit largely depends upon the nature of electrolyte employed. The electrolyte from which complex ions can be obtained, such as cyanides, provides a smooth deposit. Factor # 2. Current Density Electrodeposition depends upon the rate at which crystals grow and the rate at which fresh nuclei are formed, therefore, at low current densities the deposits are coarse and crystalline in nature. The deposit of metal will be uniform and fine-grained if the current density is used at rate higher than that at which the nuclei are formed. In case the rate of formation of nuclei is very high due to very high current density, there is a chance that the limiting value of the electrolyte is exceeded. At such instances, the deposit will be spongy and porous. Factor # 3. Temperature: A low temperature of the solution favours formation of small crystals of metal; and a high temperature, large crystals. In some cases this is very marked, a difference of only 15°C resulting in a 50% decrease in strength of the metal deposited. On the other hand, high temperature may give beneficial results due to (a) increased solubility of the salts, permitting greater metal concentration and higher current densities; (b) increased conductivity, which also permits higher

		current densities and reduces the tendency to form trees; (c) decreased occlusion of hydrogen in the deposited metal, which in many cases is the case of bad deposits. Since both (a) and (b) tend to decrease crystal size, they may in some cases counteract the tendency of temperature alone to increase the crystal size. Factor # 4. Conductivity: The use of a solution of good conductivity is important from the standpoint of view of economy in power consumption and also because it reduces the tendency to form trees and rough deposits.
5	Questions	1. What is conductivity ? 2. What is current density?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

UTILIZATION OF ELECTRICAL ENERGY & TRACTION
OMR-1

TIME- 30 MINS **TOTAL MARK- 40**

1. An electrolyte is a substance which :

a) conducts electricity

b) decomposes on heating

c) is acidic in nature

d) when dissolved in water, dissociates into ions

2. Conductivity of aqueous solution of an electrolyte depends on:

a) molecular mass of the electrolyte

b) boiling point of solvent

c) degree of ionisation

d) volume of the solvent

3. Strong electrolytes are those which :

a) dissolve readily in water

b) conducts electricity

c) dissociate into ions even at high concentration

d) dissociate into ions at high dilution

4. The electrochemical cell stops working after some time because :

a) electrode potential of both the electrodes becomes zero

b) electrode potential of both the electrodes becomes equal

c) one of the elctrodes is eaten away

d) the cell reaction gets reversed

5. In electrolysis oxidation takes place at :

a) both the electrodes b) cathode c) anode d) in the solution

6. A solution of Cu(II) sulphate is reacted with KC1 and KI. In which case will the Cu2+ be

reduced to Cu+

a) In both the cases

b) When reacted with KC1

c) When reacted with KI

d) In both the cases but in presence of H+

7. Which is not true for a standard hydrogen electrode ?

a) The hydrogen ion concentration is 1 M

b) Temperature is 25oC

c) Pressure of hydrogen is 1 atmosphere

d) It contains a metallic conductor which does not absorb hydrogen

8. The half-cell reaction is the one that :

a) takes place at one electrode

b) consumes half a unit of electricity

c) involves half a mole of electrolyte

d) goes half way to completion

9. In an electrochemical cell, anode and cathode are :

a) positively and negatively charged ions

b) positively and negatively charged electrodes

c) negatively and positively charged electrodes

d) negatively and negatively charged ions

10. More electronegative elements have :

a) negative reduction potential

b) tendency to lose electrons

c) positive reduction potential

d) positive oxidation potential

11. The strong oxidising agent has :

a) high value of reduction potential

b) high value of oxidation potential

c) low value of reduction potential

d) high tendency to lose electrons

12. The passage of electicity in the Daniell cell when Zn and Cu electrodes are connected is from :

a) Cu to Zn in the cell

b) Cu to Zn outside the cell

c) Zn to Cu outside the cell

d) Zn to Cu in the cell

13. When a lead storage battery is charged :

a) PbO_2 dissolves

b) the lead electrode becomes coated with lead sulphate

c) sulphate acid is regenerated

d) the amount of acid decreases

14. An example of a simple fuel cell is :

a) lead storage battery b) $H_2 - O_2$ cell c) Daniell cell d) Lechlanche cell

15. The number of coulombs required for the deposition of 107.87 g of silver is :

a) 48250 b) 10000 c) 96500 d) 19300

16. In the electrochemical reaction,

$2Fe^{3+} + Zn$ $Zn^{2+} + 2Fe^{2+}$

increasing the concentration of Fe^{2+} :

a) increases the cell emf b) increases the current flow

c) decreases the cell emf d) alters the pH of the solution

17. In a galvanic cell, the electrons flow from :

a) anode to cathode through the solution

b) cathode to anode through the solution

c) anode to cathode through the external circuit

d) cathode to anode through the external circuit

18. The highest electrical conductivity of the following aqueous solutions is of :

a) 0.1 M acetic acid b) 0.1 M chloroacetic acid

c) 0.1 M fluoroacetic acid d) 0.1 M difluoroacetic acid

19. Two solutions have the ratio of their concentrations 0.4 and ratio of their conductivities 0.216.

The ratio of their molar conductivities will be

a) 0.54 b) 11.574 c) 0.0864 d) 1.852

20. Which of the following is not correct ?

a) Gibbs energy is an extensive property

b) Electrode potential or cell potential is an intensive property

c) Electrical work = - Δ G

d) If half reaction is multiplied by a numerical factor, the corresponding Eo value is also multiplied by the same factor.

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Lecture Plan-09

Subject -Utilization of electric power and electric traction Semester- 5th		
1	Topic	Factors governing the electro deposition
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. What is electrolytic concentration 2. What is additional agent 3. What is throwing power
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	Factor # 5. Electrolytic Concentration: Higher current density, which is necessary to obtain uniform and fine-grain deposit, can be achieved by increasing the concentration of the electrolyte. Factor # 6. Additional Agents: The addition of acids or other substances to the electrolyte reduces its resistance, as already mentioned. There is another class of additional agents which takes little or no direct part in the chemical reactions but influences the nature of deposit, sometimes even making an otherwise unworkable process into one of practical importance. Such additional agents are glue, gums, dextrose, dextrin, gelatin, agar, alkaloids, albumen, phenol, glycerin, sugar, glucose, rubber etc. The crystal nuclei absorb the additional agent added in the electrolyte. This prevents it to have large growth and thus deposition will be fine-grained. For obtaining satisfactory deposit of zinc from zinc sulphate solution addition of glucose or certain types of sugar is necessary. Factor # 7. Throwing Power: This is the ability of electrolyte to produce uniform deposit on an article of irregular shape and is one of the most important characteristics of plating or deposition bath. The distance between the various portions of cathode and anode will be different due to irregular shape of the cathode. Due to unequal distance, the resistance of the current path through the electrolyte for various portions of the cathode will be different but the potential difference between the anode and any point on the article to be plated (cathode) will, of course be the same and the

		result will be that the current density will be more on the portion nearer to anode and it will cause uneven deposit of the metal. Throwing power can be improved in two ways—firstly by increasing the distance between the anode and cathode and secondly by reducing the voltage drop at the cathode surface. In some cases decrease of current density causes a decrease in voltage drop at cathode, leaving more voltage available for overcoming the resistance of the electrolyte, thus tending to counteract any change in current concentration. This is the reason that solutions of the cyanides of metals usually have a better throwing power than solutions of the sulphates. Factor # 8. Polarization: The rate of deposition of metal increases with the increase in electroplating current density up to a certain limit after which electrolyte surrounding the base metal becomes so much depleted of metal ions that the increase in current density does not cause increase in rate of deposition. Use of current density beyond this limit causes electrolysis of water and hydrogen liberation on the cathode. This hydrogen evolved on the cathode blankets the base metal which reduces the rate of metal deposition. This phenomenon is called the polarization. Blanketing effect can be reduced by agitating the electrolyte. With reverse current electroplating, in which at regular intervals plating current is reversed for a second or so, sufficient electron concentration is established around the base metal and the polarization effect becomes negligible even with very high overall speed of plating. The other advantages of reverse current plating are: (i) unsound and inferior metal is depleted during reverse current period and flat level surfaces are produced,
5	Questions	1. Explain throwing power ? 2. Explain electrolytic concentration? 3. What is additional agent?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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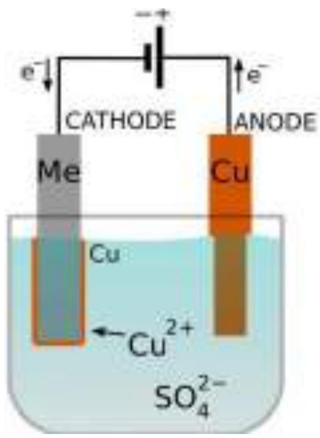
Lecture Plan-10

Subject -Utilization of Electrical Energy and Traction Semester- 5th		
1	Topic	Application of Electrolysis
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: What is application of Electrolysis
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	Applications of Electrolysis Electrolytic Refining of Metals The process of electrolytic refining of metals is used to extract impurities from crude metals. Here in this process, a block of crude metal is used as anode, a diluted salt of that metal is used as electrolyte and plates of that pure metal is used as cathode. Electrolytic Refining of Copper For understanding the process of electrolytic refining of metals, we will discuss

about an example of electrolytic refining of copper. Copper extracted from its ore, known as blister copper, is 98 to 99 % pure but it can easily be made up to 99.95% pure for electrical application by the process ofelectrorefining.

In this process of electrolysis, we use a block of impure copper as anode or positive electrode, copper sulfate acidified with sulfuric acid, as electrolyte and pure copper plates coated with graphite, as cathode or negative electrode.

The copper sulfate splits into positive copper ion (Cu+ +) and negative sulfate ion (SO4 − −). The positive copper ion (Cu+ +) or cations will move towards negative electrode made of pure copper where it takes electrons from cathode, and becomes Cu atom and is deposited on the graphite surface of the cathode.



On the other hand, the SO4 − − will move towards positive electrode or anode where it will receive electrons from anode and become radical SO4 but as radical SO4 cannot exist alone, it will attack copper of anode and form CuSO4. This CuSO4 will then dissolve and split in the solution as positive copper ion (Cu+ +) and negative sulfate ion (SO4 − −). These positive copper ions (Cu+ +) will then move towards negative electrode where it takes electrons from cathode, and become Cu atoms and are deposited on the graphite surface of the cathode. In this way, the copper of impure crude will be transferred and deposited on the graphite surface of the cathode.

The metallic impurities of anode are also merged with SO4, forming metallic sulfate and dissolve in the electrolyte solution. The impurities like silver and gold, which are not effected by sulfuric acid-copper sulfate solution, will settle down as the anode sludge or mud. At a regular interval of electrolytic refining of copper, the deposited copper is stripped out from the cathode and anode & is replaced by a new block of crude copper.

NB :- In the process of electrolytic refining of metals or simply electro refining, the cathode is coated by graphite so that the chemical deposited, can be easily stripped off. This is one of the very common applications of electrolysis.

refining of metals or simply electro refining, the cathode is coated by graphite so that the chemical deposited, can be easily stripped off. This is one of the very common applications of electrolysis.

Electroplating

The process of electroplating is theoretically same as electrorefining – only difference is that, in place of graphite coated cathode we have to place an object on which the electroplating has to be done. Let’s take an example of brass key which is to be copper-platted by using copper electroplating.

Copper Electroplating

We have already stated that copper sulfate splits into positive copper ion (Cu^{++}) and negative sulfate ion (SO_4^{--}) in its solution. For copper electroplating, we use copper sulfate solution as electrolyte, pure copper as anode and an object (a brass key) as cathode. The pure copper rod is connected with positive terminal and the brass key is connected with negative terminal of a battery. While these copper rod and key are immersed into copper-sulfate solution, the copper rod will behave as anode and the key will behave as cathode. As the cathode or the brass key is connected with negative terminal of battery, it will attract the positive cations or Cu^{++} ions and on reaching of Cu^{++} ions on the surface of the brass key, they will receive electrons from it, become neutral copper atom and are about to be deposited on the surface of the brass key as uniform layer. The sulfate or SO_4^{--} ions move to the anode and extract copper from it into the solution as mentioned in the process of electro-refining. For proper and uniform copper plating, the object (here it is brass key) is being rotated slowly into the solution.

Electroforming

Reproduction of objects by electro-deposition on some sort of mould is known as electroforming. This is another very useful example among many applications of electrolysis. For that, first we have to take the impression of objects on wax or on other wax like material. The surface of the wax mold which bears exact impression of the object, is coated with graphite powder in order to make it conducting. Then the mold is dipped into the electrolyte solution as cathode. During electrolysis process, the electrolyte metal will be deposited on the graphite coated impressed surface of the mold. After obtaining a layer of desired thickness, the article is removed and the wax is melted to get the reproduced object in form of metal shell. A popular use of electroforming is reproduction of gramophone record dices. The original recording is done on a record of wax composition. This wax mold is then coated with gold powder to make it conducting. Then this mold

		is dipped into a blue vitriol electrolyte as cathode. The solution is kept saturated by using a copper anode. The copper electroforming on the wax mold produces master plate which is used to stamp a large number of shellac discs.
5	Questions	1. Write the application of Electrolysis?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

S.K.D.A.V Govt. Polytechnic, Rourkela

Lecture Plan-11

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	ELECTRICAL HEATING: Advantages of electrical heating
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. What is advantages of electrical heating
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	**Advantages of Electric Heating** Although there are a variety of heating solutions available, not all methods of heating are created equal. Depending on the location and layout of the home being heated, some methods of heating may be better suited than others, but the advantages of electric heating usually make it the most efficient, affordable, and reliable option. At Expert Electric, we have a wealth of experience with electric heating and we can help you to find and install the best heating solution for your home. Reasons to Have Electrical Heating Installed in Your Home There are many benefits that make electric heating the way to go for most homes. Some of the best reasons to change your home to an electric heating system

		include: **Electric Heating is Safe and Reliable** With no burning materials, combustible gases, or noxious fumes, electric heating is the safest possible heating solution for your home. Most methods of electrical heating have no moving parts, which means that there is less chance of the system breaking down, so your heating system will last for a long time. **Electric Heating is Affordable and Efficient** Unlike gas and oil heating, electric heating costs have remained stable over the years. Electricity is currently the cheapest option for heating and is predicted to remain so for many years. Electric heaters can be independently wired from room to room, which means that, if a room is not in use, it does not need to be heated, which can help a home to save energy. Most electric heaters are inexpensive to install and, since electrical heating systems last longer than other methods of heating, repair and replacement costs are minimal. **Electric Heating is Environmentally Friendly** Because electric heaters do not rely on burning anything, no carbon dioxide is released into the atmosphere. Electric heating also preserves natural resources, as it does not require fuel. It generates no emissions and eliminates the risk of gas leaks. Electric Heating from Expert Electric The certified electricians from Expert Electric have experience installing a variety of different electric heating systems. Every heating system installed is optimized for the corresponding home so that heat is dispersed as efficiently as possible. Some of the electrical heating options that we install include baseboard heaters, convection heaters, electric furnaces, and electric boilers. Let us help you to find the perfect heating solution for your home.
5	Questions	1. Write the advantages of electric heating?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-12

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	Mode of heat transfer and Stephen's Law.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to:
3	Tools used	Chalk, Blackboard
4	Lecture Description	HEAT • Heat: It is denoted by symbol Q and is defined as follows "heat is something which appears at the boundary when a system changes its state due to a difference in temperature between the system and surroundings." • Heat appears only at the boundary while the change takes place inside the system. • Sign Convention: If heat flows from system to surroundings, the quantity is said

to be positive and if heat flows from surroundings to system it is said to be negative. In other words, Heat received by system=+Q Heat rejected by system=-Q

2. Difference between Heat Transfer and Thermodynamics: • Let us take an example of a hot steel bar kept in a water bath. • Thermodynamics predicts only the equilibrium temperature and state of the system but it doesn't predicts the time taken by the system to reach that equilibrium and the temperature of the hot steel bar. • Heat Transfer on the other hand helps in predicting the temperature of both the bar and the water as a function of time. • MODES OF HEAT TRANSFER: There are three modes of heat transfer: • Conduction • Convection • Radiation • Heat transfer occurs as a result of combinations of these modes of heat transfer. Heat always flows in the direction of lower temperature.

3. CONDUCTION • The transfer of heat from one part of a substance to another part of the same substance, or from one substance to another in physical contact with it, without any displacement of molecules forming the substance. • In solids, the heat is conducted by the following two mechanisms: i. Lattice Vibrations (the faster moving molecules or atoms in the hottest part of a body transfer heat impact some of their energy to adjacent molecules). ii. By transport of free electrons (Free electrons provide an energy flux in the direction of decreasing temperature). • In liquids, the process is similar but as they are more closely placed than gases, the intermolecular forces comes into play.

4. • In gases, the kinetic energy of a molecules is a function of temperature. The molecules are in constant random motion with energy and momentum. When a molecule from high T region coincides with a molecule of low T region, it loses energy by collisions.

5. • FOURIER'S LAW OF HEAT CONDUCTION: •It states that," for a homogeneous solid, the rate of heat flow is directly proportional to area of section at right angles to the direction of heat flow and to change of temperature with respect to length of the path of heat flow."

6. • Mathematically, Q a A.(dt/dx) Where, Q= Heat flow through the body per unit time (Watts) A= Surface area of heat flow (m2) dt= Temperature difference of the faces of block (K or ^{o}C) dx= Thickness of body in direction of flow (m) Thus, Q= −□.A(dt/dx) Where k= constant of proportionality also known as thermal conductivity of body.

7. • -ve sign is to take care of the decreasing temperature along the direction of increasing thickness. The temperature gradient (dt/dx) is always negative along positive x direction. • Assumptions of Fourier Law: □Conduction of heat takes place under steady state conditions.□The heat flow is unidirectional.□The temperature gradient is constant and the temperature flow is linear. □There is

no heat generation. The material is homogeneous and isotropic.

8. • Essential Features of Fourier Law: It is applicable to all matter. (Solid, Liquid, Gas) It is based on experimental evidence. It is a vector expression indicating the heat flow rate is in the direction of decreasing temperature. It helps to define the thermal conductivity of medium through which heat is conducted.

9. • The thermal conductivity of materials is defined as," amount of energy conducted through a unit area and unit thickness in unit time when the difference in temperature between the faces causing heat flow is unit temperature." • Conduction of heat occurs mostly in pure metals, less in alloys and much less in non-metals. • Thermal conductivity depends on the following factors Material Structure. Moisture Content. Density of material. Pressure and temperature of operating conditions. Units of k are W/mK or W/m^oC

10. • Thermal conductivity of a metal varies when it is heated or treated with mechanical process. • Thermal conductivity of most metals decreases with the increasing temperature. • The dependence of thermal conductivity (k) on temperature for most materials is mostly linear.

11. CONVECTION • It is the mode of energy transfer between a solid surface and the adjacent liquid or gas in motion and it involves the combined effect of conduction and fluid motion. The faster the fluid motion, the greater is the convection. • The rate equation for the convective heat transfer between a surface and an adjacent fluid is described by Newton's Law of Cooling.

12. Statement of Newton's Law of Cooling: •The coefficient of convective heat transfer (h) is defined as "the amount of heat transmitted for a unit temperature difference between the fluid and unit area of surface in unit time."

13. •Q=h.A(ts-tf) Where Q= rate of conductive heat transfer A= area exposed to heat transfer ts=Surface Temperature tf=Fluid Temperature h=coefficient of convective heat transfer Units of h are W/m2K or $W/m2^oC$

14. The value of 'h' depends on: 1. Thermodynamic and transport properties 2. Nature of fluid flow 3. Geometry of Surface 4. Prevailing thermal conditions.

15. RADIATION •It is the transfer of heat through space or matter by means other than conduction or convection. •Radiant energy(being electromagnetic radiation) requires no medium for propagation and will pass through vaccum.

16. • LAWS OF RADIATION: Wien's Law: It states that the wavelength λ corresponding to the maximum energy is inversely proportional to absolute temperature T of hot body. λmT=constant (or) λma / Stefan-Boltzmann Law: The emissive power of black body is directly proportional to fourth power of absolute temperature. Qa 4

		17. Mathematically, Q = F. □.A (T1 4 – T2 4) Where, F= a factor depending on geometry and surface properties. □ = □□□□□□ – □□□□□□□□□ □□□□□□□□= 5.67×10^{-8} W/ m2K4 A = Area of surface, m 2 **Stefan's Law:** Stefan's Law states that the radiated power density (W/m2) of a black body is proportional to its absolute temperature T raised to the fourth power...... For a true black body R = 0 and e = 1 or total absorbtion! (σ = 5.66e-8 W/m2-K4 = Stephan-Boltzmann constant).
5	Questions	1. Stefan's law? 2. Write Fourier law of heat conduction?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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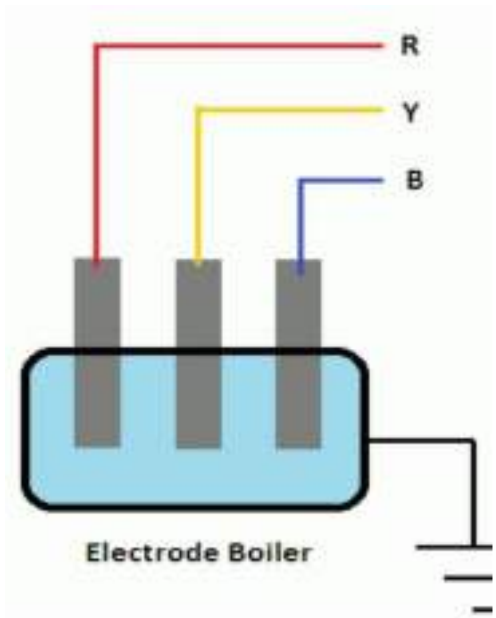
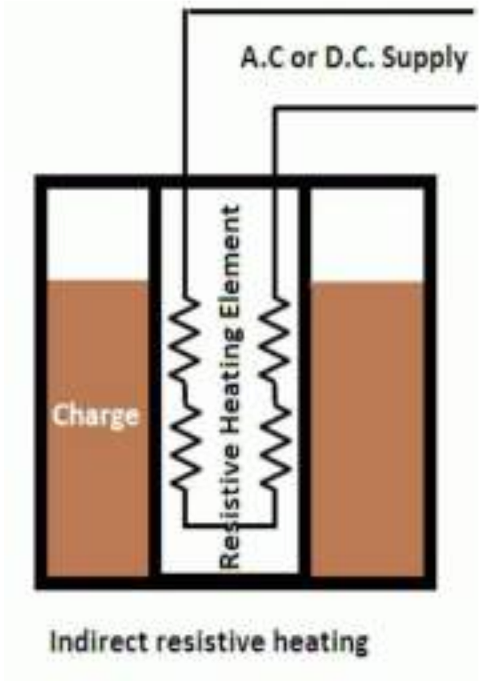
Lecture Plan-13

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	Principle of Resistance heating. (Direct resistance and indirect resistance heating.)
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. What is resistance heating

		2. What is direct and indirect resistance heating
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	The basic principle of resistance heating is that, when current passes through some resistance, power loss takes place which appears in the form of heat. There are two types of resistance heating, namely, direct resistance heating and indirect resistance heating. **Direct Resistance Heating** In the direct resistance heating, the electric current passes through the charge itself. This current produces I2R losses in the form of heat within the body itself. This principle is used in heating water by means of an electrode boiler. In the case of the electrode boiler, the electrodes are lowered into the tank filled with water. The current flow through electrodes into the water and water gets heated by I2R losses. For temperatures up to 100°C, mild steel electrodes are used.  The automatic stirring action is produced in the charge to be heated, and no external method of stirring is required to get uniform heating. Indirect Resistance Heating In the indirect resistance heating, thecurrent does not flow through the body to be heated, but it flows through the resistance elements which get heated up. The heat is then transferred from the heating element to the charge mainly by radiation or convection.

		 The figure shows a simple sketch of an indirect resistance heating oven. In this method, heat does not pass through the charge; hence natural stirring action is not there. So some external stirring apparatus is employed to get uniform heating in indirect resistance heating method. The appliances that work on this principle include immersion rod, electric kettles, electric iron etc. Temperature Control of Resistance Heating Temperature control is necessary for resistance furnace or oven. The temperature may be kept constant or varied according to requirements. The control may be manual or automatic. The following ways can control the heat in an oven or furnace: 1. Use of variable Number of Elements: In this method, the number of heating elements in working is changed, so total power input is changed. 2. Change of Connections: In this method, connections of heating elements are changed by switches at different instants according to the requirements.

		3. Transformer Tappings: We can control the voltage of the oven by using an autotransformer. It is an economical method of temperature control. Moreover, in this method, automatic temperature control can be obtained by thermostats. 4. SCRs can obtain very accurate heat control. 5. Thermostat control circuit: This is used in the automatic heating appliances to control the temperature. It is connected in series with the resistance element. It is made up of a bimetallic strip. A bimetallic strip consists of two strips of different metals, having a different coefficient of expansions, securely fastened together. When the thermostat is connected in series with the resistance element current starts flowing through the thermostat, bimetallic strip starts bending. direct resistance heating and indirect resistance heating After some time contact point A shown in figure opens and the current flowing through the resistance element stops. Now no current is flowing through the circuit, bi-metallic strip starts cooling, bi-metallic strip regains its original shape and hence contact point A closes. Thus, thermostat controls heating of element by making and breaking the circuit again and again. A screw is provided on the thermostat. It varies the tension between the two strips and hence alters the making and breaking time of the contact point A. Properties of Resistance Heating Elements High resistivity: It should have high specific resistance so that a small quantity of wire is required to produce a certain amount of heat. The low-temperature coefficient of resistance: The resistance heating element material should possess the low-temperature coefficient of resistance, so that resistance may not vary with the change in temperature.

		High melting point: The melting point of the material used should be very high so that high temperature can be obtained. Free from oxidation: It should be free from oxidation to ensure a long life of the heating element. The most common material used for the heating element is either alloy of nickel and chromium or the alloy of nickel, chromium and iron. These are used for temperatures up to 1100oC. The composition of the alloy for resistance heating elements depends upon the working temperature. For higher temperatures silicon carbide, tungsten and graphite are used. Eureka is a copper-nickel alloy approximately 56% copper and 44% nickel. It is used where the low-temperature coefficient is required, such as in instrument coils, shunts, etc.
5	Questions	1. Explain Direct resistance heating? 2. Explain indirect resistance heating?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-14

UNIT DISCUSSION- I

1. Define electrical heating.
2. Give atleast two advantages of electrical heating.

3. What are various types of electrical heating?
4. Define resistance heating.
5. State desirable properties of resistance heating element.
6. The resistance heating element is commonly made of.....
7. What are the causes of failure of heating element?
8. What is infrared heating?
9. Name few applications of infrared heating?
10. What is thermostat ?
11. What is principle of arc production?
12. What is arc heating ?
13. What is indirect arc heating ?
14. Electrods used in arc furnaces are made of.... ?
15. What is induction heating ?
16. Give two applications of induction heating ?
17. Induction furnace requires which frequency supply?
18. State application of ajax watt ?
19. Define Power factor?
20. What is dielectric heating?
21. What is an electrolyte?
22. What is an electrolyte process?
23. What is an electrodeposition?
24. What do you mean by electro chemical equivalent of a substance?
25. State a few advantages of electric heating over other forms of heating.
26. What are the specific advantages of dielectric heating?
27. Make a neat sketch of Ajax-Wyatt core type induction furnace.
28. Explain its principle of working.
29. State how the pinch effect is overcome.
30. Describe the construction and working of any type of induction furnace.
31. Describe the construction and working of any type of induction furnace.
32. Discuss the factors which determine the choice of frequency for core-less induction furnace.Enumerate the various methods of providing power supply for such a furnace.
33. Explain the process of dielectric heating.
34. What is the range of voltage and frequency used in dielectric heating and explain the reason of your answer.
35. What is the purpose of using reactors in electric arc furnaces?
36. Why are indirect arc furnaces not built in large sizes?

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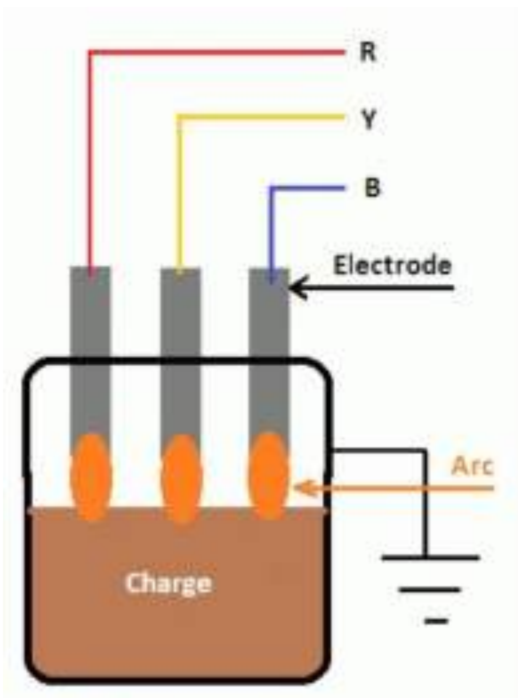
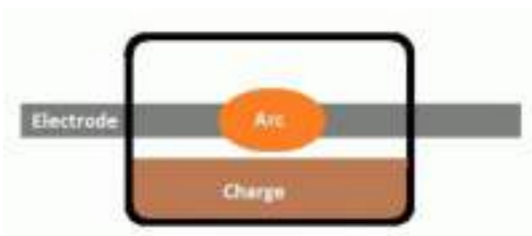
Lecture Plan-15

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	Discuss working principle of direct arc furnace and indirect arc furnace
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Explain working principle of direct arc furnace 2. Explain working principle of indirect arc furnace
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	**Working of Direct Electric Arc Furnace** The diagram shows a direct electric arc furnace. The chamber of the furnace is lined with refractory material. The arc is struck between the electrode and the charge. Three electrodes made of carbon or graphite are projected from the top of the furnace, and three-phase supply is given.  Direct Arc Furnace The current passes through them via the charge. Since the arc is in direct contact with the charge, so it is possible to produce the highest temperatures by direct electric arc furnace. As the arc passes through the charge, it will produce automatic stirring action. The arc has a negative resistance characteristic i.e.; resistance falls with the increase in temperature. Thus some current limiting device is required in the

		circuit to prevent short circuits. It may be in the form of a reactor. The direct arc furnace is very commonly used for the production of steel. There is a charging door from where the charge is supplied. And also there is an outlet for molten metal. The usual size of the direct electric arc furnace is 5 to 10 tons capacity. For melting 1 ton of steel, 1000 units of energy are consumed. **Indirect Arc Furnace Working** The figure shows an indirect arc furnace. In this furnace, the arc is formed between the two electrodes. And heat so produced is transmitted to the charge by radiation. The temperature is lower than the direct arc furnace. So indirect arc furnaces are suitable for melting metals having lower melting points e.g., non-ferrous metals such as brass, copper, zinc, etc.  Indirect Arc furnace The arc is struck between the electrodes, so only two electrodes are required. The supply is, therefore, a single phase. Since during the process of heating, the electrodes are consumed, so the feeding of electrodes to the indirect arc furnace is automatic. The furnace is cylindrical. Since the arc does not come in contact with the charge, so the automatic stirring action, which is present in the direct arc furnace, is absent. The furnace may be equipped with automatic rocking equipment. The power factor varies from 0.7 to 0.8. The electrode material and electrical equipment of indirect-arc furnace are similar to that used for direct arc furnace.
5	Questions	1. What is the principle of direct arc furnace? 2. What is the principle of indirect arc furnace?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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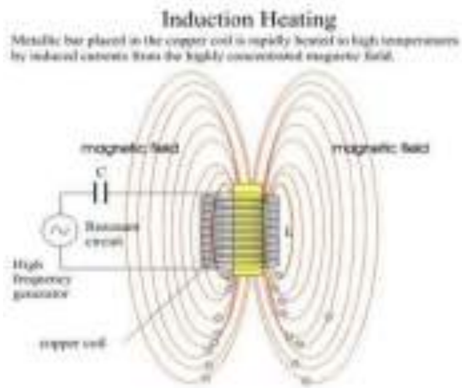
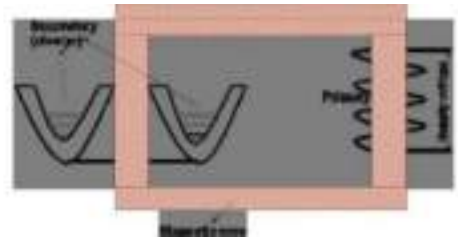
Lecture Plan-16

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION

Semester- 5th		
1	Topic	Principle of Induction heating. 1. Working principle of direct core type, vertical core type and indirect core type Induction furnace. 2. Principle of coreless induction furnace and skin effect.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. What is induction heating 2. Working principle of direct core type induction heating 3. Working principle of coreless induction furnace
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	Induction heating:  Induction heating processes make use of currents induced by electromagnetic action in the material to be heated. Induction heating is based on the principle of transformers. There is a primary winding through which an a.c current is passed. The coil is magnetically coupled with the metal to be heated which acts as secondary. An electric current is induced in this metal when the a.c current is passed through the primary coil. The following are different types of induction furnaces 1. Core type furnaces a. Direct core type induction furnace b. Vertical core type induction furnace c. Indirect core type induction furnace 2. Core less type furnaces Direct core type: The direct core type induction furnace is shown ion fig.

It consist of an iron core, crucible and primary winding connected to an a.c supply.
The charge is kept in the cruicible, which forms a single turn short circuited secondary circuit.



The current in the charge is very high in the order of several thousand amperes. The charge is magnetically coupled to the primary winding.

The change is melted because of high current induced in it. When there is no molten metal, no current will flow in the secondary.

To start the furnace molten metal is poured in the oven from the previous charge.
This type of furnace has the following drawbacks:
The magnetic coupling between the primary and secondary is very weak, therefore the leakage reactance is very high. This causes low power factor.
Low frequency supply is necessary because normal frequency causes turbulence of the charge.
If current density exceeds about 5 amps/mm2 the electromagnetic force produced by this current density causes interruption of secondary current.
Hence the heating of the metal is interrupted. It is called pinch effect.
The crucible for the charge id of odd shape and inconvenient from the metallurgical point of view.
The furnace cannot function if the secondary circuit is open.

It must be closed. For starting the furnace either molten metal is poured into the crucible or sufficient molten metal is allowed to remain in the crucible from the previous operation.

Such furnace is not suitable for intermittent services.
Indirect core type induction furnace:

I n this type of furnace induction principle has been used for heating metals.

In such furnace an inductively heated element is made to transfer its heat to the change

When the primary winding is connected to the supply, current is induced in the secondary of the metal container.

So heat is produced due to induced current. This heat is transmitted to the charge by radiation.

The portion AB of the magnetic circuit is made up of a special alloy and is kept inside the chamber of the furnace.

The special alloy will loose its magnetic properties at a particular temperature and the magnetic properties are regained when the alloy will cooled.

As soon as the furnace attains the critical temperature the reluctance of the magnetic circuit increases many times and the inductive effect correspondingly decreases thereby

cutting off the heat supply.

The bar AB is removable type and can be replaced by other, having different critical temperature. Thus the temperature of the furnace can be controlled very effectively.

Coreless induction furnace:

Coreless induction furnace also operates on the principle of transformer. In this furnace there is no core and thus the flux density will be low.

Hence for compensating the low flux density, the current supplied to the primary should have sufficiently high frequency.

The flux set up by the primary winding produces eddy currents in the charge. The heating effect of the eddy currents melts the charge.

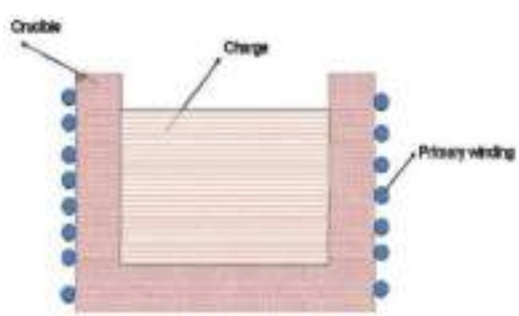
Stirring of the metals takes place by the action of the electromagnetic forces. Coreless furnace may be having conducting or non conducting containers.

Fig shows a coreless induction furnace in which container is made up of conducting material.

The container acts as secondary winding and the charge can have either conducting or non conducting properties.
Thus the container forms a short circuited single turn secondary. Hence heavy current induced in it and produce heat.

The flux produced by the primary winding produces eddy currents in the charge. The heating effects of the eddy currents melt the charge.

Stirring action in the metals takes place by the action of the electromagnetic forces.



Advantages:
Time taken to reach the melting temperature is less.

		i. Accurate power control is possible. ii. Any shape of crucible can be used. iii. The eddy currents in the charge results in automatic stirring. iv. Absence of dirt, smoke, noise, etc. v. Erection cost is less. **Applications of Induction Heating** I. Induction furnace II. Induction welding III. Induction cooking IV. Induction brazing V. Induction sealing VI. Heating to fit VII. Heat treatment **Advantages of Induction Heating** I. Optimized Consistency II. Maximized Productivity III. Improved Product Quality IV. Extended Fixture Life V. Environmentally Sound VI. Reduced Energy Consumption
5	Questions	1. What is induction heating? 2. What is working principle of direct core type induction heating? 3.what is coreless induction heating?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-17

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	Principle of dielectric heating and its application
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1.wwhat is the applications of dielectric heating
3	Tools used	Chalk, Blackboard
4	Lecture Description	Dielectric Heating Dielectric Heating is a process of electric heating by which the temperature of a dielectric (non-conducting) material is raised by the application of an alternating electric field (high voltage ac signal). The increase in temperature results in heating the substance which is in contact with the external field. Dielectric heating is sometimes called high frequency or radio-frequency heating,

		capacitive heating. This process allows uniform heating of non-metallic materials which are unable to conduct electricity. **Principle of Dielectric Heating** The principle of operation of a dielectric heater is such that a non-conducting material is present between two electrodes and an external electric field is applied across these two electrodes. Basically, a wide range of frequency is provided to the electrodes. • It is to be noted here that radio-frequency radiation is a form of energy and not a form of heat. So, it requires material matter for the conversion of heat from energy. The dielectric material which is present between the two electrodes can be anything such as wood, plastic, glass, etc. Though it is considered that a dielectric does not allow the flow of electric current through it, practically, it is not possible. So, whenever these materials are provided with a high voltage alternating supply then even minute motion of charged particles results in the flow of current which leads to dielectric losses. This resultantly produces heat within the material. Applications of Dielectric Heating **The various applications of dielectric heating are as follows:** 1. Food processing: In the field of food processing, it is used for various applications such as concentrating liquids within bottles, food cooking without outer shell removal, defrosting, dehydrating, germicidal heating, etc. 2. Preheating of plastic preforms: It is one of the significant applications of dielectric heating as no other method can perform this in a uniform manner. The raw plastic material in the form of biscuits or tablets are called plastic preform and to convert a bulk of these biscuits or tablets into a specific shape they are kept inside the required mould. 3. Basically, to get them in desired shape uniform heating up to a certain level is required before putting them in the mould. 4. Sterilization: This process suits sterilizing medical equipment and aiding items like bandages, cotton, scissors, and other gauge instruments. 5. Diathermy: To generate a specific body temperature in order to cure certain kinds of pains or diseases, body tissues and bones are subjected to dielectric heating. 6. Electronic Sewing: It is the process by which the plastic sheets of umbrellas, raincoats, medicine containers can be sealed or joined. The materials with plastic films are not joined by ordinary stitching thus by the application of heat, sealing is provided to the material under the presence of mechanical pressure.
5	Questions	1. Define dielectric heating? 2. State the application of dielectric heating?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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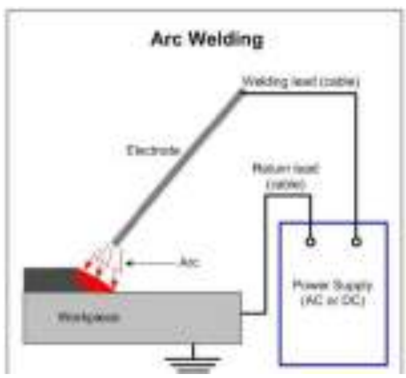
Lecture Plan-18

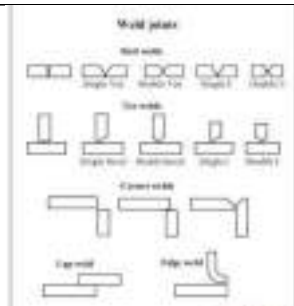
<table>
<tr><td colspan="3">Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION
Semester- 5th</td></tr>
<tr><td>1</td><td>Topic</td><td>Principle of microwave heating and it's application</td></tr>
<tr><td>2</td><td>Intended Learning Outcome</td><td>On successful completion of the Lecture students are able to:
1. What is microwave heating</td></tr>
<tr><td>3</td><td>Tools used</td><td>Chalk, Blackboard and PPT</td></tr>
<tr><td>4</td><td>Lecture Description</td><td>Principle of Microwave Heating
When an electric field is applied to metal, for example, the flow of electrons (that is, current) does not occur when a so-called insulator is placed within that electrical field for an electric conductor having freely moving electrons. However, the phenomenon of polarization, where positive and minus electric charges are displaced from the equilibrium point, resulting in a separation of the charges, does occur. Substances with this kind of nature are called a "dielectric." As frequency increases, the component electrons of a dielectric spin, collide, vibrate, rub against each other, and otherwise move violently. Changes in polarity at this time are intense, occurring several ten to several hundred million times per second. This energy becomes "heat", which causes heat to be generated inside the dielectric
Various types of application
There are many different types of applicators for devices that apply microwaves. Select the optimum method best suited to the heated substance parameters, including the following:
1. Shape (size, thickness)
2. Electrical characteristics (specific conductivity), loss factor (εr・tan δ), water content
3. State (liquid, solid, powder, etc.)
4. Purpose (preheat, heat treatment, drying, bubbling, vulcanization, etc.)
5. Treatment mode (fixed or continuous)
6. Pressure control (vacuumed or pressurized)</td></tr>
<tr><td>5</td><td>Questions</td><td>1. What is microwave heating?
2. Write the application of microwave?.</td></tr>
<tr><td>6</td><td>Reference</td><td>https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg
Utilisation of electric power and electric traction by G.C. Garg</td></tr>
</table>

S.K.D.A.V Govt. Polytechnic, Rourkela

Lecture Plan-19

<table>
<tr><td colspan="3">Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION
Semester- 5th</td></tr>
<tr><td>1</td><td>Topic</td><td>PRINCIPLES OF ARC WELDING: Explain principle of arc welding</td></tr>
<tr><td>2</td><td>Intended Learning Outcome</td><td>On successful completion of the Lecture students are able to:
1. What is arc welding
2.Explain the working principle of arc welding</td></tr>
</table>

3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	**Principles of arc welding** Arc welding is a welding process, in which heat is generated by an electric arc struck between an electrode and the work piece. Electric arc is luminous electrical discharge between two electrodes through ionized gas. **Any arc welding method is based on an electric circuit consisting of the following parts:** Power supply (AC or DC); • Welding electrode; • Work piece; • Welding leads (electric cables) connecting the electrode and work piece to the power supply.  Electric arc between the electrode and work piece closes the electric circuit. The arc temperature may reach 10000°F (5500°C), which is sufficient for fusion the work piece edges and joining them. When a long join is required the arc is moved along the joint line. The front edge of the weld pool melts the welded surfaces when the rear edge of the weld pool solidifies forming the joint. Types of weld joints are shown in the figure:

		 When a filler metal is required for better bonding, filling rod (wire) is used either as outside material fed to the arc region or as consumable welding electrode, which melts and fills the weld pool. Chemical compositions of filler metal is similar to that of work piece. Molten metal in the weld pool is chemically active and it reacts with the surrounding atmosphere. As a result weld may be contaminated by oxide and nitride inclusions deteriorating its mechanical properties. Neutral shielding gases (argon, helium) and/or shielding fluxes are used for protection of the weld pool from atmospheric contamination. Shields are supplied to the weld zone in form of a flux coating of the electrode or in other forms.
5	Questions	1. Write a application of arc welding?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-20

UTILIZATION OF ELECTRICAL ENERGY & TRACTION

OMR-2

TIME- 30 MINS **TOTAL MARK- 40**

1. Heat transfer by condition will not occur when
(A) Bodies are kept in vacuum
(B) Bodies are immersed in water
(C) Bodies are exposed to thermal radiations
(D) Temperatures of the two bodies are identical

2. Which of the following insulating materials was suitable for low temperature applications?
(A) Asbestos paper
(B) Diatomaceous earth
(C) 80 percent magnesia
(D) Cork

3. By which of the following methods the temperature inside arcane can be varied?
(A) By disconnecting some of the heating elements
(B) By varying the operating voltage
(C) By varying the current through heating elements
(D) By any of the above method

4. In an electric room heat convector the method of heating used is
(A) Arc heating
(B) Resistance heating
(C) Induction heating
(D) Dielectric heating

5. For intermittent work which of the following furnaces is suitable?
(A) Indirect arc furnace
(B) Core less furnace
(C) Either of the above
(D) None of the above
Correct Answer
6. In arc furnace the function of choke is

(A) To stabilize the arc
(B) To improve power factor
(C) To reduce severity of the surge
(D) None of the above

7. In heating the ferromagnetic material by induction heating, heat is produced due to
(A) Induced current flow through the charge
(B) Hysteresis loss taking place below Curie temperature
(C) Due to hysteresis loss as well as eddy current loss taking place in the charge
(D) None of the above factors

8. Which of the following is an advantage of eddy current heating?
(A) The amount of heat generated can be controlled accurately
(B) Heat at very high rate can be generated
(C) The area of the surface over which heat is produced can be accurately controlled
(D) All of the above

9. _________has the highest value of thermal conductivity.
(A) Copper
(B) Aluminium
(C) Brass
(D) Steel

10. When a body reflects entire radiation incident on it, then it is known as
(A) White body
(B) Grey body
(C) Black body
(D) Transparent body

11.Radiations from a black body are proportional to

A. T
B. T2

C. T3

D. T4

12. Properly of low temperature co-eflficient of heating element is desired due to which of the following reasons?

A. To avoid initial rush of current

B. To avoid change in kW rating with temperature

C. Both (A) and (B)

D. Either (A) or (B)

13.The material of the heating element for a furnace should have

A. Lower melting point

B. Higher temperature coefficient

C. High specific resistance

D. All of the above

14. In induction heating___________is abnormally high.

A. Phase angle

B. Frequency

C. Current

D. Voltage

1 5. For the transmission of heat from one body to another

A. Temperature of the two bodies must be different

B. Both bodies must be solids

C. Both bodies must be in contact

D. At least one of the bodies must have some source of heating

16. When a body reflects entire radiation incident on it, then it is known as

A. White body

B. Grey body
C. Black body

D. Transparent body

17. Heat transfer by condition will not occur when

A. Bodies are kept in vacuum

B. Bodies are immersed in water

C. Bodies are exposed to thermal radiations

D. Temperatures of the two bodies are identical

18. A perfect black body is one that

A. Transmits all incident radiations

B. Absorbs all incident radiations

C. Reflects all incident radiations

D. Absorbs, reflects and transmits all incident radiations

19. For heating element high resistivity material is chosen to

A. Reduce the length of heating element

B. Increase the life of the heating element

C. Reduce the effect of oxidation

D. Produce large amount of heat

20. Which of the following heating methods has maximum power factor?

A. Arc heating

B. Dielectric heating

C. Induction heating

D. Resistance heating

S.K.D.A.V Govt. Polytechnic, Rourkela

Lecture Plan-21

<table>
<tr><td colspan="3">Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION
Semester- 5th</td></tr>
<tr><td>1</td><td>Topic</td><td>Discuss D. C. & A. C. Arc phenomena.</td></tr>
<tr><td>2</td><td>Intended Learning Outcome</td><td>On successful completion of the Lecture students are able to:
1. Types of welding equipment
2. What is dc welding and ac welding</td></tr>
<tr><td>3</td><td>Tools used</td><td>Chalk, Blackboard and PPT</td></tr>
<tr><td>4</td><td>Lecture Description</td><td>Electric arc welding equipments falls into two natural sections

(a) Those using D.C. at the arc or D.C. Welding Equipments.

(b)Those using A.C. at the are or A.C. Welding Equipments.

(a) D.C. Welding Equipments:

It consists of mainly a motor generator set. The motor is a squirrel cage induction motor, and generator is differentially com pound to give drooping characteristics. In differential compound generator the terminal voltage falls automatically with the increase in load current.

If supply from existing system is used for welding then a Ballast resistance is put</td></tr>
</table>

		in series with the equipments and control is obtained by variation of the external resistance. This method is also suitable when a number of operators are working on the same supply system. In such cases each operator is provided with separate ballast. **(b) A.C. Welding Equipments:** A transformer reduces the voltage from that of the supply drawn to about 100 V. To regulate the current and produce the dropping characteristics required a resistance and reactance may be used. When a reactance is used it must be designed to operate well below the saturation-point of its magnetic circuit to prevent the introduc tion of harmonics which tend to lengthen the zero current period in each half cycle and allow cooling of the arc; a reactance with a air gap in the magnetic circuit is generally employed.
5	Questions	1. Explain the dc arc welding equipment? 2. Explain the ac arc welding equipment?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-22

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	D.C. & A. C. arc welding plants of single and multi-operation type.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1.
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	
5	Questions	1.
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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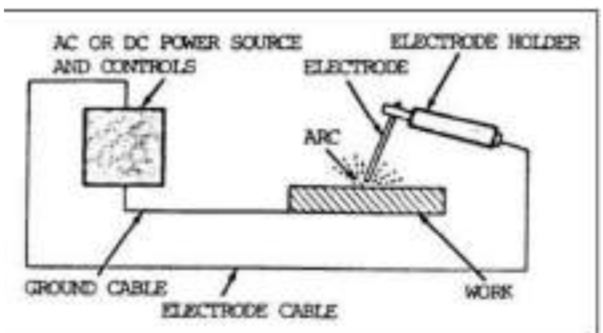
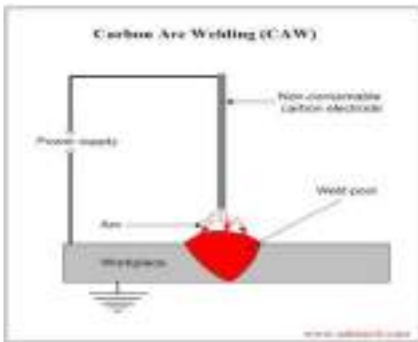
Lecture Plan-23

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	Types of arc welding.
2	Intended	On successful completion of the Lecture students are able to:

	Learning Outcome	1. Different types of arc welding
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	**Different types of arc welding:** **1. Metallic arc welding**. In the metallic arc system a metal rod is used as an electrode and the arc is struck between this electrode and work, which respectively form the two terminal. A.C. supply from the secondary of the welding transformer or D.C. supply from the D.C. welding generator depending upon whether A.C. or D.C. is used, is connected across the job and the electrode. The work is then suddenly touched by the electrode and then separated from it a little. This results in an arc between the job and the electrode. Due to the heat generated by the arc, a little portion of the work melts as also the tip of the electrode. The two pieces to be welded fuse together and when the electrode is removed the metal cools and solidifies giving a strongly welded joint. The arc drawn between the metal electrode and the work creates a temperature of over 3500°C in a small area concentrated at the point of weld and at the end ofelectrode.  The soundest welds are produced with a short arc, the potential. difference across the arc being 20 to 25 volt. In no case must it exceed 30 volt, as welds made under these conditions will not be sound. The correct current to be used in any particular job cannot be stated as, owing to the cooling effect of the different masses of metal where the weld is made, it is necessary for different current values to be used on the same thickness of sheet, depending upon the type of joint made. A lap joint as shown in the figure will require more current than the butt joint in the same material owing to the greater dissipation of heat by the greater thickness of metal under the arc where the sheet metal over-laps. **2. Carbon arc welding**. This method is normally used for welding copper and its alloys. The carbon electrode which is kept negative with respect to the work if D.C. is used. If the

carbon electrode is made positive, the carbon particles have a tendency to go into the welded joint and cause brittleness. The electrode should, therefore, be kept negative and the work positive. For this type of welding only D.C. can be used. The heat from the arc forms a molten pool and the extra metal required to make the weld is supplied by a filler rod of the same composition as that of the molten metal. Two methods of carbon arc welding are used. In one method no flux is used and in the other method flux, either in the form of powder or paste is used to prevent the weld from oxidation. Former method is limited to non-ferrous metals and the later method is usually applied for ferrous metals.



3. Atomic hydrogen arc welding.

The essentials of the atomic hydrogen arc welding process are (i) electrodes energy is supplied to an arc between two tungsten electrodes where it is transformed into heat. The source of power could be either D.C. or A.C. but A.C. supply should be chosen because it is more commonly available: (ii)

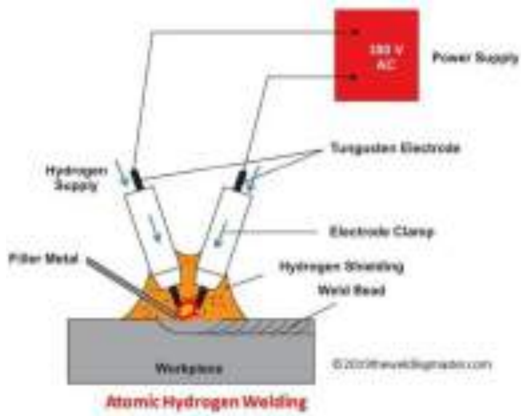
molecular hydrogen is blown through this are and transformed catalytically into the atomic form which acts as a vehicle for transfer of energy from the are to the work, and (iii) in the direction away from the arc a sudden decrease of temperature (such as that obtained at the relatively cold surface of the weld area) causes the rapid decrease in the concentration of atomic hydrogen and a release of the heat of recombination.

To strike and maintain the arc, an open circuit voltage of 300 volt is necessary and for hand welding a current range upto 50 amp.is required. This power is obtained from the transformer having a number of primary tappings to allow for various supply voltages,and a tapped reactor to permit adjustment of the current.

In this type of welding the arc is struck between two tungsten electrodes and hydrogen is passed through the arc. Due to the high temperature (about 4000°C of the arc, hydrogen changes to its atomic form. When the atomic hydrogen travels to cooler regions in the vicinity of the arc it regains its molecular form. In doing so, it gives up the energy which it had received from the arc. Thus a very intense heat is generated which is used to melt the job to be welded.

After cooling the two parts of the job are welded together. A filler rods may be used if additional metal is needed for making a joint.

This method is successfully used for welding stainless steel and moist non-ferrous metals etc.



4. Helium or argon arc welding.

This method is used for welding aluminium alloys, magnesium and magnesium alloys. An arc is struck between electrode of tungsten and the work and helium or argon is used to give an inert atmosphere so that oxidation of the welded joint does not take place.

Standard D.C. or A.C. welding equipment may be used provided the open circuit voltage is around 100 volt for A.C. and 70 volt for D.C. For welding aluminium, and their alloys stainless and high alloy steels, nickel alloys and copper alloys upto 7/8" or 1/3 cm A.C. is suitable. D.C, may be used for all other common metals, and it is essential for the welding of copper and stainless steel and alloys over 1/2" thick.

		
5	Questions	1. Explain metallic arc welding? 2. Explain carbon arc welding and atomic hydrogen arc welding?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-24

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	Explain principles of resistance welding.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Principle of resistance welding
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	Principle of Resistance Welding The name “resistance” welding derives from the fact that the resistance of the workpieces and electrodes are used in combination or contrast to generate the heat at their interface. Heat is generated by the passage of electrical current through a resistance circuit. Heat is generated in localized area which is enough to heat the metal to sufficient temperature so that the parts can be joined with the application of pressure. The force applied before, during and after the current flow forces the heated parts together so that coalescence will occur. Pressure is required throughout the entire welding cycle to assure a continuous electrical circuit through the work. Pressure is applied through electrodes. The pressure is

applied by mechanical, hydraulic or pneumatic systems.

In resistance welding, the parts to be joined are heated to plastic state by their resistance to the flow of electric current and mechanical pressure is applied to complete the weld. In this process, there are two copper electrodes in a circuit of low resistance as shown in Figure 2.1. When the current is passed through electrodes, the electrical resistance at the metal joints becomes very high. So, the metals are brought to red-hot plastic condition. Now, the mechanical pressure is applied to complete the weld. The heat developed by the current is proportional to the electric resistance of the weld.

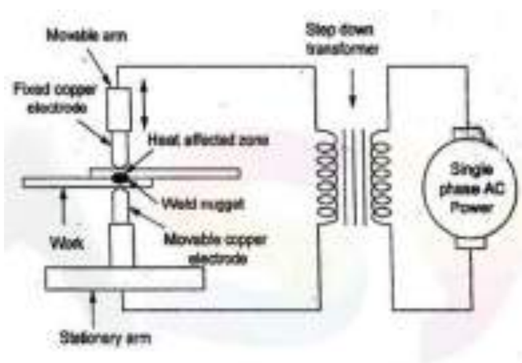

Figure 2.1 Principle of resistance welding

AC with suitable transformer is used for the power supply. Usually, 4 V to 12 V is used dependent on the composition, area and thickness of the metal to be welded. The power supply ranges from 6 to 18 kw per cm3 area used.

Resistance welding processes differ from other welding processes in the aspect that no fluxes are used and filler metal rarely used. All resistance welding operations are automatic. Therefore, all process variables are preset and maintained constant. Resistance welding equipment utilizes programmers for controlling current, time cycles, pressure and movement. Welding programs for resistance welding can become quite complex. In view of this, quality welds do not depend on welding operator skill but more on the proper set up and adjustment of the equipment and adherence to weld schedules. Three major factors involved in weld quality are as follows:

- The amount of current that passes through the work
- The pressure that the electrodes transfer to the work
- The time the current flows through the work.

The important characteristics of the resistance welding process is the transfer of heat to two parts being welded for obtaining a proper fusion even if the plates are

		dissimilar from the stand point of material or thickness. If the proper heat balance is existed only, the proper fusion can be obtained by providing an electrode with a smaller contact area at the thinner sheet and a thicker electrode at the thicker sheet together with very high current densities for short times. If two dissimilar metals with different electrical conductivities or thermal conductivities are to be joined, 1. Large contact area electrode should be used for the one which has higher electrical conductivity, and 2. Small contact area electrode should be used for the one which has higher thermal conductivity.
5	Questions	Explain the principle of resistance welding?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-25

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	Descriptive study of different resistance welding methods.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Types of resistance welding
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	**Types of Resistance Welding** There are many resistance welding processes with different uses, such as spot welding, seam welding, and butt welding. Each one has a different specific welding application that makes it optimal for a particular situation. **1. Resistance Spot Welding** Resistance spot welding has been used extensively in the automotive industry for the joining of steel and in the aerospace industry for airframe components made from aluminium alloys. It is one of the oldest and simplest forms of resistance welding, in which a weld nugget is produced by passing an electric current between the two metal components whilst they are held together between electrodes, typically made from copper-based alloys due to its superior

conductive properties

2. Resistance Seam Welding

Resistance seam welding is a variation of the standard spot welding form, however instead of spot one nugget, a series of overlapping nuggets is produced. This is usually done by replacing the conventional spot welding electrodes with wheels, which turn as the parts are fed between them. The process thus produces a continuous welded seam rather than a single spot. Seam welding is often used in the production of thin sheet, leak-tight containers such as fuel tanks, and is generally unsuitable for welding aluminium.

3. Resistance Projection Welding

Resistance projection welding is a form of resistance welding where the electricity, force, and weld time are concentrated on raised ‘projections’ across the surface. Projection welding is generally used for welds using thicker materials than the thinner metal pieces that spot welding is usually required for, and often not metals. It is used primarily in the electrical, automotive and construction industries.3. Resistance Projection Welding

Resistance projection welding is a form of resistance welding where the electricity, force, and weld time are concentrated on raised ‘projections’ across the surface. Projection welding is generally used for welds using thicker materials than the thinner metal pieces that spot welding is usually required for, and often not metals. It is used primarily in the electrical, automotive and construction industries.

4. Resistance Butt Welding

Resistance butt welding is a process in which the two components of similar cross section can be joined together in one operation that takes place simultaneously across the entirety of the affected object, rather than just in small spots. The welding application of butt welding is often in wires and rods with small diameter measurements, generally up to about 16mm diameter.

5. Flash Butt Welding

		Flash butt welding is similar, but in this case the energy transfer is primarily provided by the resistance heat arising from the parts themselves. This is a faster type of resistance welding where the welder joins the parts by applying some pressure, then by passing a heavy current through the joint which burns away surface irregularities. After the weld has generated enough heat, the parts are connected by applying heat and pressure simultaneously. This produces a forge butt weld with no melted metal remaining in the joint. • tem due to loss of orthogonality of PN codes.
5	Questions	1. What is resistance spot welding? 2. What is seam welding? 3. What is resistance butt welding?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-26

Subject-UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester-5th		
1	Topic	CLASS TEST
2	Intended learning Outcome	On successful completion of the lecture students are able to:
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	
5	Question	
6	Reference	

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Lecture Plan-27

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	**ILLUMINATION:**Nature of Radiation and its spectrum
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. What is nature of radiation
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	**The Nature of Radiation** The question, what light is has puzzled mankind since time immemorial. The ancient Greeks thought of light as consisting of a stream of extremely minute corpuscles which were shot from a hot body which when impinged human eye produced some sensation known as light. Corpuscles theory was generally accepted till about 1850 even though Huygen had put forth his theory of waves in the year 1690. The corpuscles theory gave a satisfactory picture of reflections because if corpuscles could be considered as tiny perfectly elastic spheres, they would bounce off a polished surface with the angle of reflection equal to angle of incidence. This theory, however, was unable to explain the phenomenon of interference, where the addition of two lights may produce total darkness. But the wave theory also had its problems. To account for the motion of the light

		waves through free space, physicist thought of a medium - the ether. Later mathematicians proved that such a hypothetical medium must have quite contradictory properties and the idea of its 'cxist ence was gradually given up. Also, the photo-electric effects could not be explained by wave theory. It was found necessary to introduce a new fundamental prin cipal which always applies when radiation is emitted or absorbed by matter. Energy is emitted and absorbed only in discrete quanta of magnitude hf, where h = Plank's constant (h=6.547 x 1027 erg-sec) and f= frequency in Hz. tam britta An extensive research in this field has revealed the existence of such discrete bundles or quanta of energy known as photons. Radiant energy is always emitted or absorbed in bundles and not continuously as envisaged in wave theory. It is to be noted that the valence electrons are responsible for the production of radiation in the visible region and these electrons in the outermost orbit are of utmost importance to the illuminating engineer. The energy of an atom is constant as long as the electron stays in a given stationary orbit and the atom is said to be in stationary state. Radiation is produced whenever an electron falls from an outer stationary orbit to one nearer the nucleus. Whenever, such a transition occurs, a single quantum (photon) of radiation is emitted. Thus the radiation from an atom is not continuous but is in the form of a quantum whenever an electron moves from one stationary orbit to another. But in practice as we deal with a large number of atoms, the radiation seems to be continuous in time though the individual atoms exit only in discrete quanta. Whenever an atom is excited i.e. energy is added to the atom is some way such as by collusion with an electron having a reasonably high velocity, the atom stays in the excited state (outermost electron goes to higher orbit) for a very short time of the order of 10 nano second and then reverts back to the normal state thereby emitting radiation energy. Each photon may be sidered to be associated with a wave which predicts how the photon will travel. By analogy to other waves we consider this fictitious wave also to have the three attributes frequency, wave length and velocity and the three quantities are related by the equation, $v=\lambda f$ where λ is the wave length and fthe frequency of the wave. The wave lengths used by illuminating engineers are very short and the units used are micron (10 metre) or Angstrom A (10-10 metre). The complete spectrum of radiant energy extends over a tremendous range of frequency as shown in Fig. 6.0. The visual region extends. from about 0.4 micron to 0.75 micron. The illumination engineer is interested in the conversion of electric energy into radiation energy of such frequencies as to be visible to the human eye and also in the conversion of radiant energy into electrical energy by photocells or thermocouples.
5	Questions	1. Explain the nature of radiation?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-28

<table>
<tr><td colspan="3">Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION
Semester- 5th</td></tr>
<tr><td>1</td><td>Topic</td><td>Terms used in Illuminations. [Lumen, Luminous intensity, Intensity of illumination,

MHCP, MSCP, MHSCP, Solid angle, Brightness, Luminous efficiency.]</td></tr>
<tr><td>2</td><td>Intended Learning Outcome</td><td>On successful completion of the Lecture students are able to:
1. What is lumen , luminous intensity
2. What is MHCP, MSCP,Solid angle</td></tr>
<tr><td>3</td><td>Tools used</td><td>Chalk, Blackboard</td></tr>
<tr><td>4</td><td>Lecture Description</td><td>

Terms Used in Illuminations

(i) Light.

Light is defined as the radiant energy from a hot body causing visual sensation upon the human aye.

(ii) Flux.

It is also known as luminous flux. It is defined as the total quantity of light energy radiated or emitted per second from a luminous body in the form of light waves. It is measured in lumens and is denoted by F or o. It is defined as the rate of luminous energy.

If Q lumens hour is the luminous output of source of light energy radiated is for thours. Then flux = Q/t lumens.

(iii) Light energy.

It is the energy obtained in visual radiations in a given time and in expressed in lumen hour and is denoted by Q.

(iv) Luminous efficiency or Radiant efficiency.

It is defined as the output in lumens per watt of the power consumed by the source of light. It is measured in lumens per wattage. If a bulb is rated at 500 watts and 250 volts has an efficiency of 13 lumens/watt.

Then total flux produced by bulb = 500 x 13 = 6500 lumens.

(v) Plane Angle.

Plane angle is subtended at a point in the same plane by two converging lines.

This angle is measured in radians or degrees.One radian is the angle subtended

</td></tr>
</table>

		at the centre of a circle by an arc whose length is equal to the radius of the circle. ¢=Arc/radius =l/r radians and one Radians =(180/π) degrees. The largest plane angle subtended at a point may be 2n radians. (**vi) Solid angle**. Solid angle is the angle generated by the line passing through the point in space and the periphery of the area. or A solid angle enclose a volume by an infinite number of lines lying on a surface and meeting at a point. It is measured in steradians and is denoted by w and \omega=\frac{Area}{(Radius)^{2}} steradians Where one steradian is the angle subtended at the centre of a sphere by an area on the surface of the sphere which is numerically equal to the square of the radius. The total plane angle subtended by circumference of a circle at the centre of circle is 2x radians. Similarly total solid angle subended at a point in space is obtained by considering point at the centre of sphere and the surface area of the sphere.
5	Questions	1. What is solid angle? 2. Write luminous efficiency?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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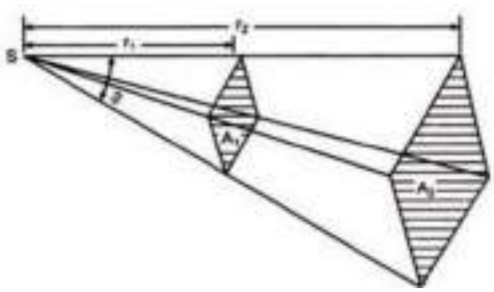
Lecture Plan-29

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	Explain the inverse square law and the cosine law.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. What is inverse square law 2. What is Lambert's cosine law
3	Tools used	Chalk, Blackboard and PPT

4	Lecture Description	**Inverse Square Law** If a source of light which emits light equally in all directions be placed at the centre of a hollow sphere, the light will fall uniformly on the inner surface of the sphere, that is to say, each square mm of the surface will receive the same amount of light. If the sphere be replaced by one of the larger radius, the same total amount of light is spread over a larger area proportional to the square of the radius. The amount which falls upon any square mm of such a surface will, therefore, diminish as the radius increases, and will be inversely proportional to the square of the distance. A similar relation holds if we have to deal with a beam of light in the form of a cone or pyramid, as shown in Fig. 7.6 (a). If we consider parallel surfaces which cut the pyramid at different distances from the source, the areas of these surfaces are proportional to the square of these distances, and, therefore, the amount of light which falls on one unit of the area of these surfaces is inversely proportional to the square of the distance from the source. This relationship is referred to as the law of inverse squares.  (a) *Inverse square law* Mathematically it can be proved as follows: Let us consider surface area A1 and surface area A2 at distances r1 and r2 respectively from the point source S of luminous intensity I and normal to the rays, as shown in Fig Let the solid angel subtended be ω steradians = I Total luminous flux radiated = Iω lumens Illumination on the surface of area A1,

$$= \frac{I\omega}{A_1} \text{ lumens/unit area}$$

and area $A_1 = \omega r_1^2$

$\therefore$ Illumination on the surface of area A_1,

$$E_1 = \frac{I\omega}{\omega r_1^2} = \frac{1}{r_1^2} \text{ lumens per unit area}$$

Similarly illumination on the surface of area A_2,

$$E_2 = \frac{I\omega}{\text{[illegible]}} = \frac{I\omega}{\text{[illegible]}} = \frac{1}{\text{[illegible]}} \text{ lumens per unit area.}$$

Hence the illumination of a surface is inversely proportional to the square of the distance between the surface and the light source provided that the distance between the surface and the source is sufficiently large so that the source can be regarded as a point source.

Lambert's Cosine Law:

Very often the illuminated surface is not normal to the direction of light as AC in Fig.(b)but is inclined as AB. The area over which the light is spread is then increased in the ratio-

$$\frac{AB}{AC} = \frac{1}{\cos\theta}$$

and the illumination decreases in the ratio $\frac{\cos\theta}{1}$.

(b) Lambert's cosine law

The expression for the illumination then becomes

$E = I\cos\theta/r^2$

According to this law the illumination at any point on a surface is proportional to the cosine of the angle between the normal at that point and the direction of luminous flux.

5	Questions	1. Explain inverse square law? 2. What is Lambert's cosine law?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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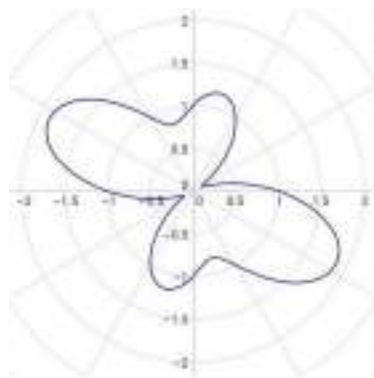
Lecture Plan-30

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th

1	Topic	Explain polar curves.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. What is polar curve
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	**Polar Curves** A polar curve is a shape constructed using the polar coordinate system. Polar curves are defined by points that are a variable distance from the origin (the pole) depending on the angle measured off the positive xx-axis. Polar curves can describe familiar Cartesian shapes such as ellipses as well as some unfamiliar shapes such as cardioids and lemniscates.  $r = 1 - \cos\theta \sin 3\theta$ Whereas Cartesian curves are useful to describe paths in terms of horizontal and vertical distances, polar curves are more useful to describe paths which are an absolute distance from a certain point. One practical use of polar curves is to describe directional microphone pickup patterns. A directional microphone will pick up different qualities of sound depending on what location the sound comes from outside of the microphone. For example, a cardioid microphone has a pickup-pattern in the shape of a cardioid.
5	Questions	1. Explain the polar curve?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-31

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	Describe light distribution and control. Explain related definitions like maintenance factor and depreciation factors
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Explain distribution of light. 2. Say how to control distribution of light.
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	**LIGHT DISTRIBUTION TYPES** There are a lot of important factors to look at when choosing the right light fixture. Lumens per watt, fixture type, warranties, and color temperatures are all super important, but how about light distribution? One of the more critical things to look at is the type of light distribution from the fixture itself. If you know how to properly figure a layout based on using the right fixture and the correct distribution of light, you can lower fixture counts and get a better lit project. You can have a great fixture, but if you're not putting light where you want, it's only as good as the layout! **DIFFERENT TYPES OF LIGHT DISTRIBUTION PATTERNS** Light distribution is basically the projected pattern of light that a fixture will disperse onto a surface. Let's look at light distribution types in relation to inside the building and outside the building. Indoor beam spread patterns Outdoor light distribution patterns **INDOOR LIGHTING** Indoor fixtures and bulbs typically used beam spreads like a 30-degree flood, 270-degree omnidirectional & 8-degree spot. The beam spread that you use is the determining factor in how far the light will travel onto a surface. If you have a wide beam of 40 degrees and you want it to pin spot on a sign 20 feet away, you are barking up the wrong tree. However, if you use an 8-degree spot, that sign will look great! **OUTDOOR LIGHTING** Outdoor light fixtures use light distribution patterns. The patterns define how light is dispersed from a fixture and is defined by the point at which 50% of the fixture's luminous intensity is met. You will see these distributions used regularly in parking lot lights, area lighting, floodlighting & roadway lighting. While there are other symmetrical and asymmetrical light distribution types, the industry has identified 5 main types of light distribution patterns:

		Type I Type I distribution is popular for lighting walkways, paths, and sidewalks. We use this type of lighting in the middle of a pathway and it is great for narrow pathways. Type II Type II is commonly used on larger walkways and in roadway lighting as it is good for a larger but still narrow area. You will see this type used on side streets or jogging paths. Type III Type III is very commonly used in roadway lighting as it gives a bit more coverage further from the point source outward. Type III lighting needs to be placed to the side of the area, allowing the light to project outward and fill the area. Type IV This light distribution produces a semicircular light that should be used on the sides of buildings and walls. This type does a great job lighting up a parking area perimeter. Type V This is the widest distribution pattern and is excellent for illuminating the inside portions of a parking lot or a 4-way intersection. With this basic knowledge, selecting the right light distribution type for your security lights or parking lot lights becomes easier. **Lighting control system** A lighting control system is an intelligent network based lighting control solution that incorporates communication between various system inputs and outputs related to lighting control with the use of one or more central computing devices. Lighting control systems are widely used on both indoor and outdoor lighting of commercial, industrial, and residential spaces. Lighting control systems serve to provide the right amount of light where and when it is needed. Lighting control systems are employed to maximize the energy savings from the

		lighting system, satisfy building codes, or comply with green building and energy conservation programs. Lighting control systems are often referred to under the term Smart Lighting. **Maintenance factor** Lighting systems are subject to contamination and wear until the end of their useful life. A lighting system's maintenance factor indicates how much of the initial luminous flux remains available at the end of its service life. The maintenance factor must be determined by the planning lighting engineer and the new value of the luminous flux multiplied by it. This ensures that the specified light intensity remains at the end of the service life. The light intensity needed for a given work activity is set out by the EN 12464-1 International Lighting Standard. A lighting system's maintenance factor is composed of four factors. **Depreciation factor:** It is defined as the ratio of initial illumination to the ultimate maintained illumination on the working plane. ....Coefficient of utilization or utilization factor: It is defined as the ratio of total number of lumens reaching the working plane to the total number of lumens emitting from source.
5	Questions	1. What is light distribution? 2. What is maintainance factor?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-32

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	Design simple lighting schemes and depreciation factor
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. Design schemes of lighting.
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	See below

Types of Lighting Schemes:

i. Direct Lighting:

It is most commonly used type of lighting scheme. In this lighting scheme more than 90 percent of total light flux is made to fall directly on the working plane with the help of deep reflectors. Though it is most efficient but causes hard shadows and glare. It is mainly used for industrial and general out-door lighting.

ii. Semi-Direct Lighting:

In this lighting scheme 60 to 90 percent of the total light flux is made to fall downwards directly with the help of semi-direct reflectors, remaining light is used to illuminate the ceiling and walls. Such a lighting system is best suited to rooms with high ceilings where a high level of uniformally distributed illumination is desirable. Glare in such units is avoided by employing diffusing globed which not only improve the brightness towards the eye but improve the efficiency of the systems with reference to working place.

iii. Semi-Indirect Lighting:

In this lighting scheme 60 to 90 percent of total light flux is thrown upwards to the ceiling for diffuse reflection and the rest reaches the working plane directly except for some absorption by the bowl. This lighting scheme is with soft shadows and glare free. It is mainly used for indoor light decoration purposes.

		iv. Indirect Lighting: In this light scheme more than 90 percent of total light flux is thrown upwards to the ceiling for diffuse re•flection by using inverted or bowl reflectors. In such a system the ceiling acts as the light source, and the glare is reduced to mini•mum. The resulting illumination is softer and more diffused, the shadows are less prominent and the appearance of the room is much improved over that which results from direct lighting. It is used for decoration purposes in cinemas theatres and hotels etc. and in workshops where large machines and other obstructions would cause trouble some shadows of direct lighting is employed. v. General Lighting: In this scheme lamps made of diffusing glass are used which give nearly equal illumination in all directions. Outdoor Lighting Scheme: (i) Street lighting (ii) Flood lighting General Ideas about Street Lighting: The main purpose of street lighting is to make the traffic and obstructions on the road clearly visible for safety point of view and convenience. The principle employed for street lighting is different from that of interior lighting. As the area is open i.e. there is no wall, the illumination level required is low. The question of colour rendering is also of minor importance. Principle of Street Lighting In fact in case of interior lighting the objects are seen by light reflected by them but in case of street lighting the objects are seen in bright back ground. The reflectors used throw the light on the road at a very large angle of incidence. It is observed that a motorist needs to see the objects about 30 m away. Fig. 2.39 shows the principle of street lighting. The observer is able to see the object through reflected light as shown. Generally an average number of 8 to 15 lumens per square meter is considered sufficient illumination on the street. Mercury vapour lamps and sodium discharge lamps has been found most economical due to lower power consumption for a given amount of light. Colour consideration does not matter much in street

		lighting. Flood Lighting: The meaning of flood lighting is flooding of large surfaces with light from powerful projectors. Purpose of Flood Lighting: Followings are the main purpose of employing flood lighting: (i) Beautification: For enhancing beauty of building at night such as public place, ancient building and monuments, religious building on important festive occasions etc. (ii) Industrial and Commercial Flood-Lighting: For illumina•tion railway yards, sports stadiums, car parks, construction site, quarries etc. For flood lighting it is necessary to concentrate the light from light source info a narrow beam. The type of reflector and its housing used for concentrating the light into narrow beam is known as flood light projector. The reflecting surface is made of silvered glass or stainless steel. Metal reflectors being more robust are usually preferred. The casing and its mounting are arranged in such a manner that the beam can be in horizontal and a vertical direction on site. When higher wattage of projector lamp i.e. 500 W or 1000 W is used in projectors, then ventilation may be provided for cooling properly.p
5	Questions	1. What is indirect lightning? 2. What is industrial and commercial flood lightning?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-33

UNIT DISCUSSION-II

1. What is the difference between illumination and light.
2. What are the advantages of artificial lighting produced electrically?
3. What are the two types of lighting used in commercial purposen lighting?
4. What is the difference between maintenance factor and depreciation factor ?

5. What are the two laws of illumination?
6. Tungsten is selected as the filament material. Why?
7. Why does the efficiency of a filament lamp increases with the rease in operating voltage?
8. Why sodium discharge lamps are not used for general lighting?
9. What is flood lighting?
10. Why thermal type starter for use in flourscent tubes have me obsolete these days.
11. What are the two types of reflecting surfaces widely used for light fittings?
12. What is the amount of luminous per square metre should be provided for difficult or fine works.
13. What type of lamps are used at stages and video studios? Ans. Halogen lamps.
14. What is the purpose of illuminating show case or window case?
15. What are the faults or accident which are likely to occur in show case or window case wiring?
16. How illumination differs from product to another?
17. State the types of lamps used for illumination of show cases:
18. What are three kinds of reflector of light?
19. Which is the most recent incandescent lamp which gives better Ilumination?
20. How much lumens per square metre should be required for adequate visual representation?
21. Define and explain the following terms in connection with illumination.
    (i) Illumination (ii) Luminous flux (iii) Coefficient of utilisation
    (iv) Depreciatifactor(v)Reflection factor
22. State the laws of illumination?
23. With the help of a diagrammatic sketch, showing connections explain the principle of working of a fluorescent lamps with particular refer ence to starter and choke. Mention the special field of application of such lamps.
24. Define the following terms:
    (i) horizontal candle power (ii)Luminous flux
25. Explain with the help of a circuit diagram the working of a fluorescent lamp.
26. What are the advantages of fluorescent lighting over plain mercury, discharge lighting?
27. Why condenser is used in the fluorescent tube circuit?
28. Why are coiled coil gas filled incandescent lamps more efficient than evacuated lamps?
29. What is welding?
30. What is the difference between plastic welding and fusion welding?
31. What is resistance welding?
32. What is the difference between electric arc welding and resistance welding?
33. Why alternating current is found most suitable for resistance welding?
34. What is meant by electrode?
35. What is the difference between carbon arc welding and metallic are welding?

36. What type of electric supply is suitable for electric arc welding?
37. What is main draw back of electron beam process?
38. In laser beam welding, the location of the focal point of the beam w.r.t. the surface of the work piece is very important. Why?
39. What properties would you seek for selection of electrode material for spot welding?
40. What is the difference between carbon arc welding and metallic arc welding?

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Lecture Plan-34

<table>
<tr><td colspan="3">Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION
Semester- 5th</td></tr>
<tr><td>1</td><td>Topic</td><td>Constructional features and working of filament lamps , effect of variation of voltage on working of filament lamps.</td></tr>
<tr><td>2</td><td>Intended Learning Outcome</td><td>On successful completion of the Lecture students are able to:
1. What is working of filament lamps</td></tr>
<tr><td>3</td><td>Tools used</td><td>Chalk, Blackboard and PPT</td></tr>
<tr><td>4</td><td>Lecture Description</td><td>This lamp consists filament of a fine wire of the metal, tungsten supported in a glass envelope and the air evacuated from the glass bulb – hence this bulb also called as vacuum lamp.

Filaments are constructed by tungsten due to its high melting point property. It operates at a high temperature of 2300o C and has an output of about 8 lm/W (Lumen/watt).

At temperatures above 2000oC, the filament does not melt, but it begins to break up and particles of that filament fly in to the side of the lamp. This causes the glass of bulb to become black. It can frequently observable in torch light bulbs.

The evaporation causes weak spots in the filament resulting in uneven resistance in filament, which sets up hot spots and it burns out and breaks.

Bulb filled with an inert gas like Argon and nitrogen which do not support combustion and they reduce the rate of evaporation.

The operating temperature of an inert gas filled lamp is about 2700oC and output is nearly 12 lm/W.

tungsten-filament-lamp

Some important points relate to Tungsten filament lamp:

There are two types of filaments are, Single coil filament and Coiled coil filament. The main advantage of a coiled coil lamp is that higher the light output.</td></tr>
</table>

		General lighting service (GLS) filament lamps used in homes have a bayonet cap (BC) fixed at end of lamp. The rated life of general lighting service (GLS) lamps is 1000 hours. The rated life of tungsten filament lamp is achieved in 'normal conditions of use'. The normal conditions of use are, operated cap up Free from vibration Voltage not excess the rated voltage. Suitable light fittings. Operating a tungsten filament lamp at a voltage higher than its rated voltage will reduce its life and Lower operating voltage will extend its life. At higher voltage, the filament lamp gives a whiter and a more bluish light and operates at brighter and higher efficiency.
5	Questions	1. Briefly explain the construction of filament lamps. 2. Write the construction of filament lamps.
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-35

Subject - UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	Discharge lamps
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1.Define discharge lamp.
3	Tools used	Chalk, Blackboard

4	Lecture Description	**The Discharge lamps are one of the reliable sources of artificial light energy, which fundamentally works by converting electrical energy to light energy by ionizing the gas particles. Discharge lamp gained popularity, in the mid-twentieth century. The best part of the discharge lamps, it can produce lights of different colors. Due to its working principle, it is cheap, long-lasting, and can be used for a number of applications. For a period, these discharge lamps are even used in aircraft and vehicles. But due to the advent of technology, they have been replaced with CFL bulbs and LEDs.** **Design Features of Discharge Lamp** **As shown in the diagram, a discharge lamp consists of one gas tube, wherein a gas at low pressure is maintained. The gas tubes are filled with noble gasses like argon, neon, krypton, etc. Additionally, they are also filled with sodium mercury amalgam.** **The sodium and mercury amalgam is required to draw an arc. At the end of the gas tube, two electrodes are placed, which produce very high voltage. The electrodes are excited by an alternating voltage source. In order to produce a high voltage, a ballast is maintained in series with the alternating voltage source. The ballast works on the inductor principle. In some advanced lamps, electronic ballast is also used.** **Symbolically a discharge lamp is shown as above. It has shown an ellipse, representing the glass tube, and two symbols at the end representing the electrodes.** **Working of Lamp** **The discharge lamps work on the fundamental principle that, light is emitted when the electrons have collided with each other. To obtain this, the first gas an ionized at very high voltage. At low pressure, gas maintained in a glass tube. At the end of the tube, two electrodes are placed which are excited with an alternating voltage source.** **When the source produces high voltage, then the gas inside the glass tube gets ionized due to the ionization of the molecules. The ionized electrons tend to move from one end to the other end. In this process, the ionized electrons collide with the unionized electrons and emit light. Based on the properties of the gas different colors of light can be produced**

Discharge Lamp Types

There are three types of discharge lamps which are given as

PCBWay

Low-pressure Discharge

In low-pressure discharge lamps, the working pressure is much less than the atmospheric pressure. Like the fluorescent lamps, and the sodium lamps. They produce 200 lumens per watt. One advantage of low-pressure discharge is they have a long life span. The low-pressure discharge lamps require the ballast to create a high voltage. Ballasts fundamentally work on the principle of inductance. The voltage across the inductor is directly proportional to the product of inductance and the rate of change of current (di/dt). With increase.

High-Pressure Discharge

The working pressure of these lamps is much higher than the atmospheric pressure. Different categories are there in high-pressure discharge lamps such as metal halide lamps, high-pressure sodium lamps, and high-pressure sodium mercury lamps.

High-intensity Discharge

The high-intensity discharge lamps are very efficient as compared to pressure based discharge lamps. They have a long life span and can save a lot amount of energy. The high-intensity lamps use an arc array to produce intense light. For this they need high voltage, i.e. they need ballasts just like fluorescent lights. The ballast helps to build up the high voltage.

OurPCB

Sodium Vapor Lamp

Sodium Vapor Lamp

For this reason, when turned on, the high-intensity lamps take up to ten minutes to reach peak value. The high-pressure mercury vapor lamp, metal halide lamps, and high-pressure sodium lamps are some of the examples of high-intensity lamps.

Lamp Examples

We shall see the high-intensity discharge lamps for this case. The mercury lamps are mostly used for street lighting due to their long span of life. They provide about 50 lumens per watt. They also provide huge energy savings while usage. The metal halide lamps are mostly used for indoor purposes like stadiums,

		outdoor areas, sports facilities, etc. **They provide very bright and intense light. They are much simple in construction as compared to mercury lamps and provide more lumens per watts. The high-pressure sodium lamps provide a warm light with a comparative high life span and more lumens per watt.** **Advantages** **The advantages of the discharge lamp are** **Long life** **Less in cost** **Can be used for a number of applications** **Fewer heat dissipations as compared to fluorescent bulbs.** **Can be designed for different colors** **Disadvantages** **The disadvantages of the discharge lamp are** **More costly as compared to fluorescent bulbs. This factor is replaced by CFL bulbs** **Applications** **The applications of the discharge lamp are** **Street Lighting** **Sport Arenas** **Aircrafts** **Gymnasiums** **Industries**
5	Questions	1. What is discharge lamp? 2. Write application of discharge lamp?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-36

Subject -Advanced Communication Engineering Semester- 6th		
1	Topic	State Basic idea about excitation in gas discharge lamps
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1.what is excitation in gas discharge lamp
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	Gas-discharge lamps are a family of artificial light sources that generate light by sending an electric discharge through an ionized gas, a plasma.  Germicidal lamps are simple low-pressure mercury vapor discharges in a fused quartz envelope. Typically, such lamps use a noble gas (argon, neon, krypton, and xenon) or a mixture of these gases. Some include additional substances, like mercury, sodium, and metal halides, which are vaporized during startup to become part of the gas mixture. In operation, some of the electrons are forced to leave the atoms of the gas near the anode by the electric field applied between the two electrodes, leaving these atoms positively ionized. The free electrons thus released flow onto the anode, while the cations thus formed are accelerated by the electric field and flow towards the cathode.

		Typically, after traveling a very short distance, the ions collide with neutral gas atoms, which transfer their electrons to the ions. The atoms which lost an electron during the collisions ionize and speed toward the cathode while the ions which gained an electron during the collisions return to a lower energy state while releasing energy in the form of photons.Light of a characteristic frequency is thus emitted. In this way, electrons are relayed through the gas from the cathode to the anode. The color of the light produced depends on the emission spectra of the atoms making up the gas, as well as the pressure of the gas, current density, and other variables. Gas discharge lamps can produce a wide range of colors. Some lamps produce ultraviolet radiation which is converted to visible light by a fluorescent coating on the inside of the lamp's glass surface. The fluorescent lamp is perhaps the best known gas-discharge lamp. Compared to incandescent lamps, gas-discharge lamps offer higher efficiency,but are more complicated to manufacture and most exhibit negative resistance, causing the resistance in the plasma to decrease as the current flow increases. Therefore, they usually require auxiliary electronic equipment such as ballasts to control current flow through the gas, preventing current runaway (arc flash). Some gas-discharge lamps also have a perceivable start-up time to achieve their full light output. Still, due to their greater efficiency, gas-discharge lamps were preferred over incandescent lights in many lighting applications, until recent improvements in LED lamp technology.
5	Questions	1. Explain excitation gas discharge lamp?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-37

UTILIZATION OF ELECTRICAL ENERGY & TRACTION

OMR-3

TIME- 30 MINS

TOTAL Mark- 40

1. Radiant efficiency of the luminous source depends on

(A) shape of the source

(B) temperature of the source

(C) wavelength of light rays

(D) all of the above.

2. Light waves travel with a velocity of

(A) 3 x 1010cm/s

(B) 3 x 1012cm/s

(C) 3 x 1015 cm/s

(D) 3 x 1018 cm/s.

3. Carbon arc lamps are commonly used in

(A) domestic lighting

(B) street lighting

(C) cinema projectors

(D) photography.

4. The unit of solid angle is

(A) solid angle

(B) radian

(C) steradian

(D) candela.

5. Candela is the unit of

(A) Luminous flux

(B) Luminous intensity

(C) Wavelength

(D) None of the above.

6. The unit of luminous flux is

(A) steradian

(B) candela

(C) lumen

(D) lux.

7. The illumination is directly proportional to the cosine of the angle made by the normal to the illuminated surface with the direction of the incident flux. Above statement is associated with

(A) Planck's law

(B) Macbeth's law of illumination

(C) Bunsen's law of illumination

(D) Lambert's cosine law.

8. Which curve represents life of the lamp ?

(A) curve A

(B) curve B

(C) curve C

(D) curve D.
9. Illumination level required for precision work is around

(A) 50 lm/m2

(B) 100 lm/m2

(C) 200 lm/m2

(D) 500 lm/m2.
10. Which of the following will need the highest level of illumination ?

(A) Proof reading

(B) Bed rooms

(C) Hospital wards

(D) Railway platforms.

11. Which of the following will need lowest level of illumination ?

(A) Displays

(B) Fine engraving

(C) Railway platform

(D) Auditoriums.
12. Which of the following lamp gives nearly monochromatic light ?

(A) Sodium vapor lamp

(B) GLS lamp

(C) Tube light

(D) Mercury vapor lamp.
13. The illumination level in houses is in the range

(A) 10-20 lumen/m2

(B) 30 - 50 lumen/m2

(C) 40-75 lumen/m2

(D) 100-140 lumen/m2.

14. Luminous efficiency of a fluorescent tube is

(A) 5- 10 lumens/watt

(B) 15-20 lumens/watt

(C) 30 - 40 lumens/watt

(D) 60 - 65 lumens/watt.

15. One lumen per square meter is the same as

(A) One lux

(B) One candela

(C) One foot candle

(D) One lumen meter.

16. Which of the following lamp gives nearly monochromatic light ?

(A) Sodium vapor lamp

(B) GLS lamp

(C) Tube light

(D) Mercury vapor lamp.

17. The illumination level in houses is in the range

(A) 10-20 lumen/m2

(B) 30 - 50 lumen/m2

(C) 40-75 lumen/m2

(D) 100-140 lumen/m2.

18. Luminous efficiency of a fluorescent tube is

(A) 5- 10 lumens/watt

(B) 15-20 lumens/watt

(C) 30 - 40 lumens/watt

(D) 60 - 65 lumens/watt.

19. One lumen per square meter is the same as

(A) One lux

(B) One candela

(C) One foot candle

(D) One lumen meter.

20. Which of the following lamp gives nearly monochromatic light ?

(A) Sodium vapor lamp

(B) GLS lamp

(C) Tube light

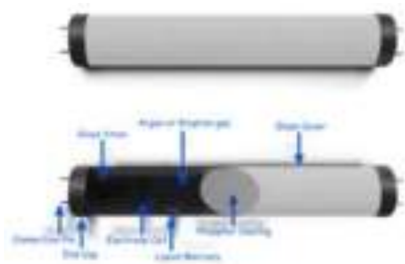
(D) Mercury vapor lamp.

Lecture Plan-38

<table>
<tr><td colspan="3">Subject -Utilization of electric power and electric traction
Semester- 5th</td></tr>
<tr><td>1</td><td>Topic</td><td>State constructional factures and operation of Fluorescent lamp. (PL and PLL Lamps)</td></tr>
<tr><td>2</td><td>Intended Learning Outcome</td><td>On successful completion of the Lecture students are able to:
1. Constructional features and operation of fluorescent lamp</td></tr>
<tr><td>3</td><td>Tools used</td><td>Chalk, Blackboard and PPT</td></tr>
<tr><td>4</td><td>Lecture Description</td><td>

What is Fluorescent Lamp?

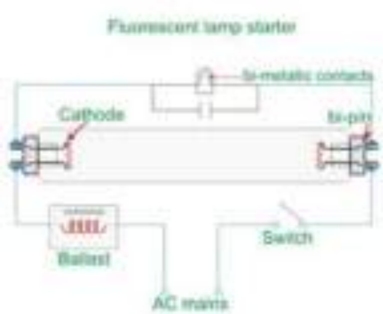
A fluorescent lamp is a low weight mercury vapour lamp that uses fluorescence to deliver visible light. An electric current in the gas energizes mercury vapor which delivers ultraviolet radiation through discharge process and the ultraviolet radiation causes the phosphor coating of the lamp inner wall to radiate visible light.



A fluorescent lamp has changed over electrical energy into useful light energy to a great deal more proficiently than incandescent lamps. The normal luminous viability of fluorescent lighting frameworks is 50 to100 lumens per watt, which is a few times the adequacy of incandescent lamps with equivalent light yield.

How does a Fluorescent Lamp work

Before going through the working principle of a fluorescent lamp, we will first show the circuit of a fluorescent lamp in other words circuit of tube light.



Here we connect one ballast, and one switch and the supply is series as shown. Then we connect the fluorescent tube and a starter across it.

</td></tr>
</table>

- When we switch ON the supply, full voltage comes across the lamp and as well as across the starter through the ballast. But at that instant, no discharge happens, i.e., no lumen output from the lamp.
- At that full voltage first the glow discharge is established in the starter. This is because the electrodes gap in the neon bulb of starter is much lesser than that of the fluorescent lamp.
- Then gas inside the starter gets ionized due to this full voltage and heats the bimetallic strip. That causes to bend the bimetallic strip to connect to the fixed contact. Now, current starts flowing through the starter. Although the ionization potential of the neon is more than that of the argon but still due to small electrode gap, a high voltage gradient appears in the neon bulb and hence glow discharge gets started first in the starter.
- As soon as the current starts flowing through the touched contacts of the neon bulb of the starter, the voltage across the neon bulb gets reduced since the current, causes a voltage drop across the inductor(ballast). At reduced or no voltage across the neon bulb of the starter, there will be no more gas discharge taking place and hence the bimetallic strip gets cool and breaks away from the fixed contact. At the time of breaking of the contacts in the neon bulb of the starter, the current gets interrupted, and hence at that moment, a large voltage surge comes across the inductor(ballast).

$$V = L\frac{di}{dt}$$

$$\text{Where, } L \text{ is inductance of inductor}$$

$$\text{and } \frac{di}{dt} \text{ is rate of change of current.}$$

- This high valued surge voltage comes across the fluorescent lamp (tube light) electrodes and strikes penning mixture (mixture argon gas and mercury vapor).
- Gas discharge process gets started and continues and hence current again gets a path to flow through the fluorescent lamp tube (tube light) itself. During discharging of penning gas mixture the resistance offered by the gas is lower than the resistance of starter.
- The discharge of mercury atoms produces ultraviolet radiation which in turn excites the phosphor powder coating to radiate visible light.
- Starter gets inactive during glowing of fluorescent lamp (tube light) because no current passes through the starter in that condition.

PL & PLL

PLUG-IN LAMPS

LED lamps are the ideal solution to replace plug-in compact fluorescent lamps and are just as easy to install. These LED lamps are available in 2-pin and 4-pin, as well as in horizontal and vertical models. They are perfect for use in commercial buildings, hotels and offices by providing an eco-friendly and cost saving alternative to any ceiling fixture, recessed can,

		downlight or sconce.
5	Questions	1. Write working principle of fluorescent lamp? 2. What is PL & PLL lamps?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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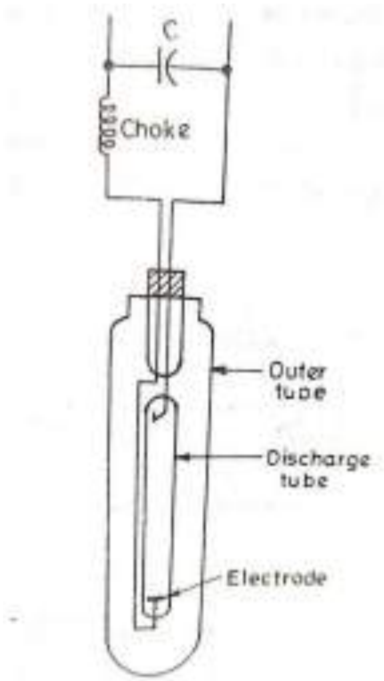
Lecture Plan-39

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	Sodium vapor lamps
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. What is sodium vapor lamps
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	Sodium Vapour Lamp A sodium lamp consists of an inner bulb of special glass containing the sodium and the inert gas either neon or argon and is fitted with two filaments. This is enclosed in a larger bulb which is evacuated to prevent the escape of heat as much as possible. The efficiency of the bulb is high as the radiations from it are the yellow sodium lines which are near the maximum of the visibility curve. The normal operating temperature of the tube is around 300°C. A small transformer is included in the circuit for heating the cathode and a choke for stabilizing the discharge. To begin with, the discharge starts in the inert gas and the temperature of the lamp

		 Sodium vapour lamp increases gradually until sufficient vapour is present to conduct the current. Since the ionization potential of the metal vapour is low, it will carry nearly all the current and the spectrum of the inert gas will practically disappear soon. Thus the bulb starts as a neon lamp and then gradually changes from red to yellow as the temperature rises. Because of the monochromatic yellow light which makes the objects appear as grey, these lamps are used for street and high way lighting.
5	Questions	1. Explain sodium vapour lamp?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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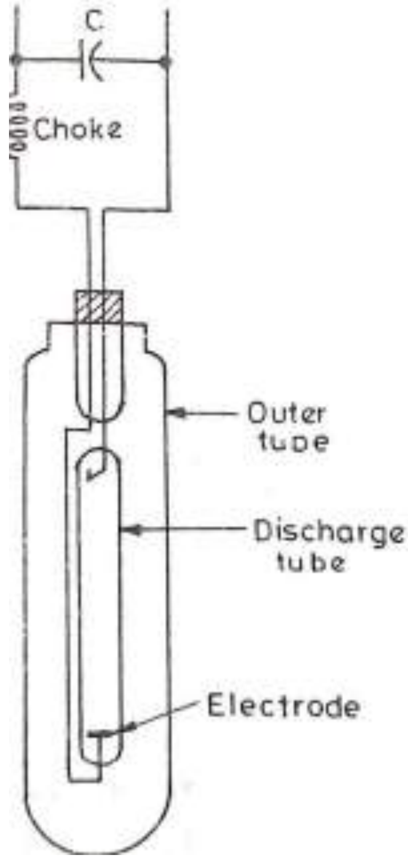
Lecture Plan-40

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	High pressure mercury vapor lamps.
2	Intended	On successful completion of the Lecture students are able to:

	Learning Outcome	1. What is mercury vapor lamps
3	Tools used	Chalk, Blackboard
4	Lecture Description	**Mercury Vapour Lamp** The mercury vapour lamp is similar in construction to the sodium vapour lamp and consists of a double glass bulb. Since the cathode is maintained in an incondenscent state by ionic bombard ment, no heating circuit is required. The operation of the mercury vapour lamp is at low mercury vapour pressure gives blue light and a high proportion of the ultra violet rays and hence are unsatisfac tory. A more recent development is a high pressure mercury lamp in which a small but carefully measured drop of mercury is intro duced in the bulb. The vapour pressure rises until all the mercury is vaporized when the light given out is with a bluish tinge. Here also an inert gas is introduced for initial discharge of the mercury. When the mercury vapour lamp is in use and say it goes out, it will not restart untill it has cooled down and the vapour pressure fallen to a value sufficiently low to allow restricking of the discharge in the inert gas. The overall efficiency of the bulb is approximately 35



Mercury vapour lamp

lumen per watt, and the p.f. 0.65. A still more recent development in mercury lamps is an ultra pressure lamp operating at about 40 atmosphere and has an efficiency comparable with the sodium vapour lamps. These lamps are used for outdoor and in dustrial lighting.

High Pressure Mercury Vapour Lamps

It was found that pressure-efficiency characteristics in a mer cury vapour discharge improved when the pressure was increased at about one atmosphere and this led to high pressure mercury lamp development. First HP mercury lamp was developed in 1906 in Germany by R Kuch and T Retschinsky. But lamps for general lighting came in early 1930's. The practical problems were to find suitable glass for

		the discharge tube, glass to metal seal and starting at line voltage. Fused silica is affected by the barium oxide used for electrode emission and therefore, the temperature is restricted to 750°C. Molybdenum foil is used at the pinch and a combination of rare gas filling at a pressure of few mm of mercury is used. The starting problem is solved by an auxiliary electrode.
5	Questions	1. Explain mercury vapor lamp ?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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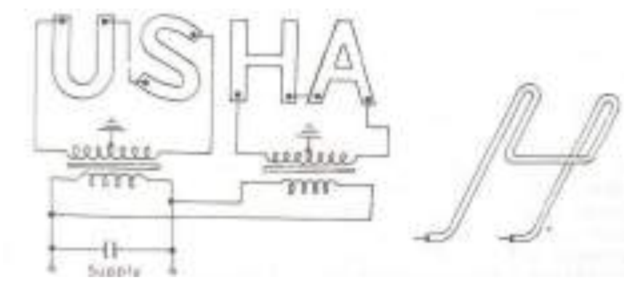
Lecture Plan-41

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	Neon sign lamps
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. What is neon sign lamp
3	Tools used	Chalk, Blackboard
4	Lecture Description	Neon Lamps. These lamps are operated at a very low temperature of about 200 that is why, these are also called cold cathode discharge lamps. The electrodes are made from iron or nickle cylinders without any coating and practically these do not emitelectrons. The electrodes are housed at the two ends of the tube which contains neon gas. Neon gas discharge gives a characteristic red light. Mixture of mercury vapour and argon gas gives bluish green colour whereas mercury and neon give blue colour. Voltage required for starting and maintaining the discharge may be as high as about 10,000 volts. The particular voltage is required for a particular length of the tube and it may be obtained from a transformer. Since the cold cathode lamps require about 1.5 to 2 times the operating voltage at the time of start, so these are provided with high reactance transformers. From the safety point of view, no part of the installation should have a voltage

		more than 5000 V above ground. Therefore, the transformers H.V. is limited to 10,000 volts and the centre point of the winding is earthed as shown in figure (a). Neon lamps are generally used for advertising. Most of the letters have two ends at which electrodes are placed. In letters having more than two ends the tube path is repeated for some portion as shown in Fig.(b) for letter 'H' used in the above advertising. If the length of the advertising sign tube requires voltage more than 10,000 volts it is split up into parts.  (a) Neon sign lamp. (b) Tube Arrangement For 'H'
5	Questions	1. Explain neon sign lamp ?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-41

Subject- UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	CLASS TEST QUESTION DISCUSSION AND DISTRIBUTION OF EVALUATED ANSWER TO THE STUDENT
2	Intended learning outcome	On successful completion of the lecture students are able to:
3	Tools used	Chalk, Blackboard
4	Lecture description	

Lecture Plan-42

<table>
<tr><td colspan="3">Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION
Semester- 5th</td></tr>
<tr><td>1</td><td>Topic</td><td>High lumen output & low consumption fluorescent lamps</td></tr>
<tr><td>2</td><td>Intended Learning Outcome</td><td>On successful completion of the Lecture students are able to:
1. what is high lumen output ?
2. What is low consumption fluorocent lamp?</td></tr>
<tr><td>3</td><td>Tools used</td><td>Chalk, Blackboard</td></tr>
<tr><td>4</td><td>Lecture Description</td><td>When you're shopping for light bulbs, compare lumens to be sure you're getting the amount of light, or level of brightness, you want. The Lighting Facts Label will help. This label makes it easy to compare bulb brightness, color, life, and estimated operating cost for the year

Buy Lumens, Not Watts

In the past, we bought light bulbs based on how much energy, or watts, they use. But now it makes more sense to buy lights based on how much light they provide.

When you're shopping for light bulbs, think about the brightness you want by comparing lumens instead of watts. A lumen is a measure of the amount of brightness of a lightbulb -- the higher the number of lumens, the brighter the lightbulb.

What's a Lumen?

Lumens measure how much light you are getting from a bulb. More lumens means it's a brighter light; fewer lumens means it's a dimmer light.

Lumens are to light what

• Pounds are to bananas
• Gallons are to milk

Lumens let you buy the amount of light you want. So when buying light bulbs, think lumens, not watts.

The brightness, or lumen levels, of the lights in your home may vary widely, so</td></tr>
</table>

		here's a rule of thumb: ★ To replace a 100 watt (W) incandescent bulb, look for a bulb that gives you about 1600 lumens. If you want something dimmer, go for fewer lumens; if you prefer brighter light, look for more lumens. ★ Replace a 75W bulb with an energy-saving bulb that gives you about 1100 lumens ★ Replace a 60W bulb with an energy-saving bulb that gives you about 800 lumens ★ Replace a 40W bulb with an energy-saving bulb that gives you about 450 lumens. A compact fluorescent lamp (CFL), also called compact fluorescent light, energy saving light and compact fluorescent tube, is a fluorescent lamp designed to replace an incandescent light bulb; some types fit into light fixtures designed for incandescent bulbs. The lamps use a tube that is curved or folded to fit into the space of an incandescent bulb, and a compact electronic ballast in the base of the lamp. Compared to general-service incandescent lamps giving the same amount of visible light, CFLs use one-fifth to one-third the electric power, and last eight to fifteen times longer. A CFL has a higher purchase price than an incandescent lamp, but can save over five times its purchase price in electricity costs over the lamp's lifetime.[1] Like all fluorescent lamps, CFLs contain toxic mercury,[2] which complicates their disposal. In many countries, governments have banned the disposal of CFLs together with regular garbage. These countries have established special collection systems for CFLs and other hazardous waste. The principle of operation remains the same as in other fluorescent lighting: electrons that are bound to mercury atoms are excited to states where they will radiate ultraviolet light as they return to a lower energy level; this emitted ultraviolet light is converted into visible light as it strikes the fluorescent coating, and into heat when absorbed by other materials such as glass. CFLs radiate a spectral power distribution that is different from that of incandescent lamps. Improved phosphor formulations have improved the perceived color of the light emitted by CFLs, such that some sources rate the best "soft white" CFLs as subjectively similar in color to standard incandescent lamps. White LED lamps compete with CFLs for high-efficiency lighting,and General Electric has stopped production of domestic CFL lamps in the United States in favour of LEDs.
5	Questions	1. Explain low consumption fluorocent lamp?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-43

<table>
<tr><td colspan="3">Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION
Semester- 5th</td></tr>
<tr><td>1</td><td>Topic</td><td>NDUSTRIAL DRIVES: State group and individual drive</td></tr>
<tr><td>2</td><td>Intended Learning Outcome</td><td>On successful completion of the Lecture students are able to:
1. Advantages of industrial drive
2. Advantages of group drive</td></tr>
<tr><td>3</td><td>Tools used</td><td>Chalk, Blackboard</td></tr>
<tr><td>4</td><td>Lecture Description</td><td>Individual Drive and Group Drive
There are generally large number of machines to be driven in any industry.
In individual drive each machine and tool is driven by a separate motor.
Although in this type initial cost is increased. In group drive a single motor of large capacity is to drive number of machines and tools in industry.
Advantages of Group Drive
Here, advantages of group drive will be the disadvantages of individual drive.
Initial Cost. Initial cost of Group drive is less as compared to thatof the individual drive.
Sequence of Operation. Group drive system is useful because all the operations are stopped simultaneously.
Space Requirement. Less space is required Group drive as compared to individual drive.
Low Maintenance Cost. It requires little maintenance as compared to individual drive.
Disadvantages of Group Drive
Although the initial cost of group drive is less but yet this system is not</td></tr>
</table>

		used extensively because of following disadvantages: (1) Power factor. Group drive has low power factor. (ii) Efficiency. Group drive system when used and if all the machines are not working together the main motor shall work at very much reduced load. (iii) Reliability. In group drive if the main motor fails whole industry will come to stand-still. (iv) Flexibility. Such arrangement is not possible in group drive i.e. this arrangement is not suitable for the place where flexibility is the prime factor. (v) Speed Group drive does not provide constant speed. (vi) Types of machines. Group drive is not suitable for driving heavy machines such as cranes, lifts and hoists etc. 2.8. Advantages of Individual Drive (1) Individual drive give desired operation as each machine is driven by its own individual motor.
5	Questions	1. Write the advantages of industrial drive? 2. Write the disadvantage of group drive?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-44

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 6th		
1	Topic	Method of choice of electric drives.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. What is electric drives
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	Choice of Electrical Drives: Choice of Electrical Drives depends on a number of factors. Some of the important factors are: Steady state operation requirements: Nature of speed torque characteristics, speed regulation, speed range, efficiency, duty cycle, quadrants of operation, speed fluctuations if any, ratings. Transient operation requirements: Values of acceleration and deceleration, starting, braking and reversing performance. Requirements related to the source: Type of source, and its capacity, magnitude of voltage, voltage fluctuations, power factor, harmonics and their effect on other loads, ability to accept regenerated power. Capital and running cost, maintenance needs, life. Space and weight restrictions if any. Environment and location. Reliability.
5	Questions	1. write the method of choice electric drives?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-45

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION

Semester- 5th		
1	Topic	Explain starting and running characteristics of DC and AC motor.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. What is starting current of dc motor 2. What is running current and there characteristics
3	Tools used	Chalk, Blackboard
4	Lecture Description	SERIES-WOUNDMOTORCHARACTERISTICS□ Series-Wound DC Motors □Speed varies automatically with the load, increasing as the load decreases□Generally limited to case where a heavy power demand is necessary to bring the machine up to speed, as in the case of certain elevator and hoist installations, etc□Series-wound motors should never be used where the motor cab be started without load, since they will race to a dangerous degree COMPOUND-WOUNDMOTORCHARACTERISTICS□Compound-Wound DC Motors□Combination of the shunt wound and series wound types - combines the characteristics of both.□Characteristics may be varied by varying the combination of the two windings. □ Generally used where severe starting conditions are met and constant speed is required at the same time. INDUCTIONMOTORCHARACTERISTICS□Squirrel-Cage Induction Motors □The most simple and reliable of all electric motors.□ Essentially a constant speed machine, which is adaptable for users under all but the most severe starting conditions. □Squirrel cage motors are common in many household appliances including washing machines, dishwashers, and stand alone fans. INDUCTIONMOTORCHARACTERISTICS□Wound-Rotor Induction motor □Used for constant speed-service requiring a heavier starting torque than is obtainable with squirrel cage type. □Because of its lower starting current, this type is frequently used instead of the squirrel-cage type in larger sizes. □These motors are also used for varying-speed-service. □Speed varies with this load, so that they should not be used where constant speed at each adjustment is required, as for machine tools. **AC MOTOR CHARACTERISTICS** Two important characteristics of AC motors that need to be considered for any application are running speed and starting torque. Running speed is dependent on the power supply frequency, the number of motor poles and the amount of slip. The frequency and number of poles define what the synchronous speed of the motor will be. Our blog post on Synchronous and Induction motors will further explain the differences between these motor types. (See Figure 1 to determine

		synchronous speed.) The amount of speed below the synchronous speed the motor runs at (expressed as a percent), is defined as the slip. AC motors require slip to induce current in the rotor, and the amount of slip changes as the load on the motor changes. In order to change the speed of an AC motor, the frequency must be changed. This is accomplished by a motor control, and the most common is a Variable Frequency Drive (VFD). CHARACTERISTIC: STARTING TORQUE Another key characteristic is the starting torque of the motor. In comparison to other motor types, starting torque is the chief limitation of an AC motor. A single phase motor will not start on and must have help. Single phase motors are defined by the methods they use to start. Some common types of single phase motors are the shaded pole motor, the split phase motor, the permanent split capacitor motor (also called the single value capacitor motor), and the two value capacitor motor. All these motor types either use an out of phase secondary coil or a capacitor to create a secondary phase to start the motor. Remember, if your application requires the motor to start with a load on it, consult your motor manufacturer to ensure the motor has enough torque to start at load and to ensure the correct motor type is specified for your application. In comparison to single phase motors, three phase motors have a higher power density, a higher starting torque, and are more efficient than single phase motors. They start on their own, eliminating the need for a starting winding or capacitor. The same speed calculation applies to three phase motors as single phase, so a VFD is required to change the motor speed. Also, when a three phase power source is unavailable, controls are capable of converting single phase power into three phase power, making the three phase motor more versatile. AC motors work great for a number of different applications such as pumps, conveyors, and commercial products. Remember the key characteristics when considering an AC motor, and consult your motor manufacturer with your specific application specifications to ensure you are getting the correct motor for your application.
5	Questions	1. Write starting and running characteristics of dc motor? 2. Write starting and running characteristics of ac motor?

6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-46

<table>
<tr><td colspan="3">Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION
Semester- 5th</td></tr>
<tr><td>1</td><td>Topic</td><td>State Application of:

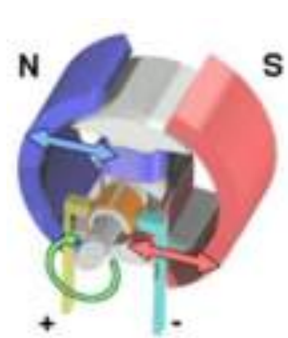
.1. DC motor.

2. 3-phase induction motor.

.3. 3 phase synchronous motors.

4. Single phase induction, series motor, universal motor and repulsion motor.</td></tr>
<tr><td>2</td><td>Intended Learning Outcome</td><td>On successful completion of the Lecture students are able to:
1. What is dc motor
2. What is 3- phase induction motor and 3-phase synchronous motor
3.</td></tr>
<tr><td>3</td><td>Tools used</td><td>Chalk, Blackboard</td></tr>
<tr><td>4</td><td>Lecture Description</td><td><u>DC MOTOR</u>

A DC motor is any of a class of rotary electrical motors that converts direct current electrical energy into mechanical energy. The most common types rely on the forces produced by magnetic fields. Nearly all types of DC motors have some internal mechanism, either electromechanical or electronic, to periodically change the direction of current in part of the motor.



Workings of a brushed electric motor with a two-pole rotor (armature) and</td></tr>
</table>

permanent magnet stator. "N" and "S" designate polarities on the inside axis faces of the magnets; the outside faces have opposite polarities. The + and - signs show where the DC current is applied to the commutator which supplies current to the armature coils

DC motors were the first form of motor widely used, as they could be powered from existing direct-current lighting power distribution systems. A DC motor's speed can be controlled over a wide range, using either a variable supply voltage or by changing the strength of current in its field windings. Small DC motors are used in tools, toys, and appliances. The universal motor can operate on direct current but is a lightweight brushed motor used for portable power tools and appliances. Larger DC motors are currently used in propulsion of electric vehicles, elevator and hoists, and in drives for steel rolling mills. The advent of power electronics has made replacement of DC motors with AC motors possible in many applications.

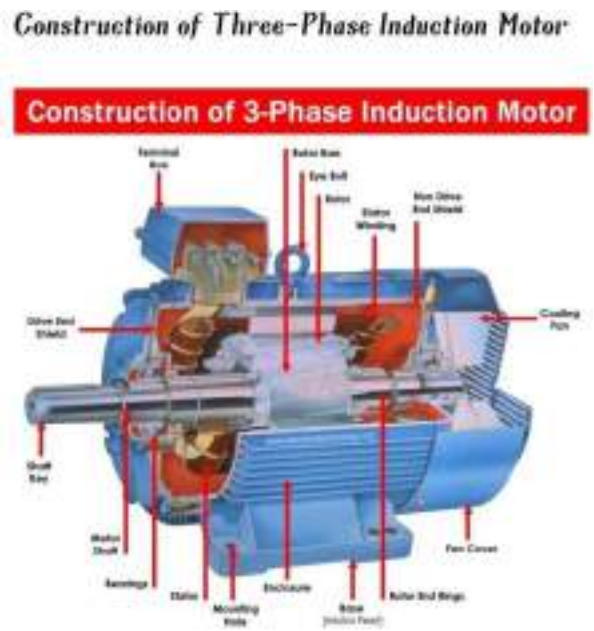
Three Phase Induction

The motor is used to convert an electrical form of energy into mechanical form. According to the type of supply, motors are classified as AC motors and DC motors. In today post, we will discuss the different types of three phase induction motors with working and applications.

The induction motor especially three phase induction motors are widely used AC motor to produce mechanical power in industrial applications. Almost 80% of the motor is a three-phase induction motor among all motors used in industries. Therefore, the induction motor is the most important motor among all other types of motor.

What is a 3-Phase Induction Motor?

A three phase induction motor is a type of AC induction motors which operates on three phase supply as compared to the single phase induction motor where single phase supply is needed to operate it. The three phase supply current produces an electromagnetic field in the stator winding which leads to generate the torque in the rotor winding of three phase induction motor having magnetic field.



The construction of an induction motor is very simple and robust. It has mainly two parts;

Stator

Rotor

Stator

As the name suggests, the stator is a stationary part of the motor. The stator of the induction motor consists of three main parts;

Stator Frame

Stator Core

Stator Winding

Stator Frame

The stator frame is the outer part of the motor. The function of the stator frame is

		to provide support to the stator core and stator winding. It provides mechanical strength to the inner parts of the motor. The frame has fins on the outer surface for heat dissipation and cooling of the motor. The frame is casted for small machines and it is fabricated for a large machine. According to the applications, the frame is made up of die-cast or fabricated steel, aluminum/ aluminum alloys, or stainless steel. Related Post: DC Machine – Construction, Working, Types and Applications Stator Core The function of the stator core is to carry the alternating magnetic flux which produces hysteresis and eddy current loss. To minimize these losses, the core is laminated by high-grade steel stampings thickness of 0.3 to 0.6 mm. These stampings are insulated from each other by varnish. All stampings stamp together in the shape of the stator core and fixed it with the stator frame. An inner layer of the stator core has a number of slots. Stator Winding The stator winding is placed inside the stator slots available inside the stator core. Three-phase winding is placed as a stator winding. And three-phase supply is given to the stator winding. The number of poles of a motor depends on the internal connection of the stator winding and it will decide the speed of the motor. If the number of poles is greater, the speed will less and if the number of poles is lesser than the speed will

high. The poles are always in pairs. Therefore, the total number of poles always an even number. The relation between synchronous speed and number poles is as shown in the below equation,

NS = 120f / P

Where;

f = Supply Frequency

P = Total Number of Poles

Ns = Synchronous Speed

As the end of winding connected to the terminal box. Hence, there are six terminals (two of each phase) in the terminal box.

According to the application and type of starting methods of motors, the stator winding is connected in star or delta and it is done by the connection of terminals in the terminal box.

Rotor

As the name suggests, the rotor is a rotating part of the motor. According to the type of rotor, the induction motor is classified as;

Squirrel Cage Induction Motor

Phase Wound (Wound Rotor) Induction motor / Slip-ring Induction Motor

The construction of the stator is same in both types of induction motors. We will discuss the types of rotors used in 3-phase induction motors in the following section of types of three phase induction motors.

Types of Three Phase Induction Motors

Three phase motors are classified mainly in two categories based on the rotor winding (Armature coil winding) i.e. squirrel cage and slip ring (wound rotor motor).

Squirrel Cage Induction Motor

Slip-ring or Wound Rotor Induction Motor

Squirrel Cage Induction Motor

The shape of this rotor is resembling the shape of the cage of a squirrel. Therefore, this motor is known as a squirrel cage induction motor.

The construction of this type of rotor is very simple and rugged. So, almost 80% of the induction motor is a squirrel cage induction motor.

The rotor consists of a cylindrical laminated core and has slots on the outer periphery. The slots are not parallel but it is skewed at some angle. It helps to prevent magnetic locking between the stator and rotor teeth. It results in smooth operation and reduces the humming noise. It increases the length of the rotor conductor due to this the rotor resistance is increased.

The squirrel cage rotor consists of rotor bars instead of the rotor winding. The rotor bars are made up of aluminum, brass, or copper.

Rotor bars are permanently shorted by end rings. So, it makes a complete close path in the rotor circuit. The rotor bars are welded or braced with the end rings to provide mechanical support.

The rotor bars are short-circuited. Therefore, it is not possible to add external resistance to the rotor circuit.

In this type of rotor, the slip rings and brushes are not used. Hence, the construction of this type of motor is simpler and more robust.

Slip-ring or Wound Rotor Induction Motor

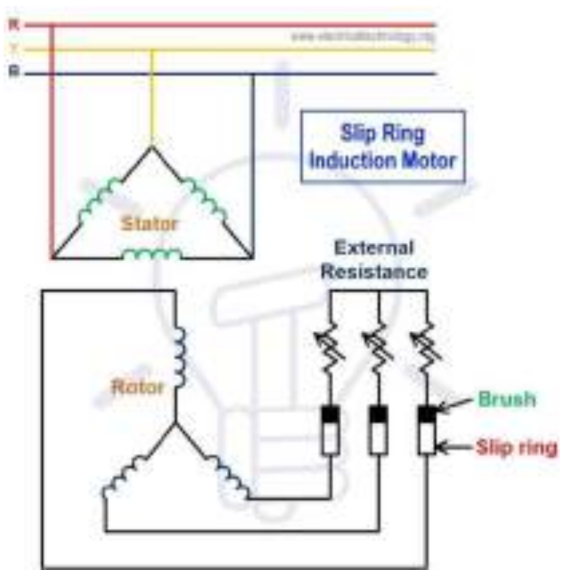
Slip-ring induction motors are also known as wound rotor motor. The rotor consists of a laminated cylindrical core with slots on the outer periphery. The rotor winding is placed inside the slots.

In this type of rotor, the rotor winding is wounded in such a way that, the number of poles of rotor winding is the same as the number of poles of the stator winding. The rotor winding can be connected as a star or delta.

End terminals of rotor windings are connected to the slip-rings. So, this motor is known as a slip-ring induction motor.

The external resistance can easily connect with the rotor circuit through the slip-ring and brushes. And it is very useful for controlling the speed of the motor and improving the starting torque of the three-phase induction motor.

An electrical diagram of slip-ring three-phase induction motor with external resistance is shown in the below figure.



The external resistance is used only for the starting purpose. If it is remains connected during the running condition, it will increase the rotor copper loss.

High rotor resistance is good for the starting condition. So, the external resistance is connected with the rotor circuit during the starting condition.

When motor running near the speed of the actual speed, the slip-rings are short-circuited by the metal collar. By this arrangement, the brushes and external resistance is removed from the rotor circuit.

This reduces the rotor copper loss as well as friction in brushes. The rotor construction is a little bit complicated compared to the squirrel cage motor due to the presence of brushes and slip-rings.

The maintenance of this motor is more. So, this motor only used when variable speed control and high starting torque are needed. Otherwise, the squirrel cage induction motor is more preferred over slip-ring induction motor.

Working Principle of Three-Phase Induction Motor

The stator winding is overlapped at 120° (electrically) to each other. When a three-phase supply is given to the stator winding, the rotating magnetic field (RMF) induced in the stator circuit.

The speed of the rotating magnetic field is known as synchronous speed (NS).

According to Faraday's law, EMF induced in the conductor due to the rate of change of flux ($d\Phi/dt$). The rotor circuit cut the stator magnetic field and an EMF induced in the rotor bar or rotor winding.

The rotor circuit is a close path. So, due to this EMF current will flow through the rotor circuit.

Now, we know that the current-carrying conductor induces the magnetic field. So, the rotor current induces a second magnetic field.

		The relative motion between the stator flux and rotor flux, the rotor starts to rotate to reduce the cause of relative motion. The rotor tries to catch the stator flux and starts rotating. **Three Phase synchronous motors** In the electrical systems, we use either in industries, power stations or domestic needs, motors and generators have become a common thing. With the demand for high energy efficient and less power consuming systems, the invention of new models of these electrical devices is seen. The basic calculating factor for motors and generators reliable operation is the Power factor. It is the ratio of applied power over the required power. Usually, the total powered consumed at the industries and factories are calculated based on the power factor. So, power factor should always be maintained at unity. But due to the rise of reactive power in these devices power factor decreases. To maintain power factor at unity many methods are introduced. The synchronous motor concept is one of them. What is Synchronous Motor? The definition of synchronous motor states that " An AC Motor in which at steady state, rotation of the shaft is in sync with the frequency of applied current". The synchronous motor works as AC motor but here the total number of rotations made by the shaft is equal to the integer multiple of the frequency of the applied current. The synchronous motor doesn't rely on induction current for working. In these motors, unlike induction motor, multiphase AC electromagnets are present on the stator, which produces a rotating magnetic- field. Here rotor is of a permanent magnet which gets synced with the rotating magnetic- field and rotates in synchronous to the frequency of current applied to it.Synchronous Motor Working Principle Working of synchronous motors depends on the interaction of the magnetic field of the stator with the magnetic field of the rotor. The stator contains 3 phase windings and is supplied with 3 phase power. Thus, stator winding produces a 3 phased rotating Magnetic- Field. DC supply is given to the rotor.

The rotor enters into the rotating Magnetic-Field produced by the stator winding and rotates in synchronization. Now, the speed of the motor depends on the frequency of the supplied current.

Speed of the synchronous motor is controlled by the frequency of the applied current. The speed of a synchronous motor can be calculated as

Ns=60f/P=120f/p

where, f = frequency of the AC current (Hz)

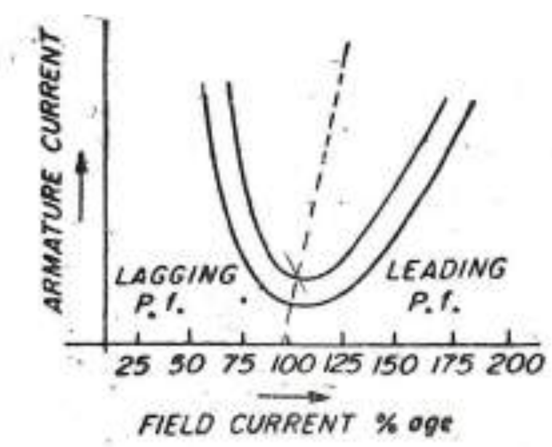
p = total number of poles per phase

P = total pair number of poles per phase.

If the load greater than breakdown load is applied, the motor gets desynchronized. The 3 phase stator winding gives the advantage of determining the direction of rotation. In case of single-phase winding, it is not possible to derive the direction of rotation and the motor can start in either of the direction. To control the direction of rotation in these synchronous motors, starting arrangements are needed.

Single Phase Induction Motor.

Its construction is similar to a polyphase induction motor except the stator is provided with 1-0 winding and



the speed torque characteristics are the same as that of a 3-4 induction motor

		with only difference that it has no self starting torque. **Repulsion Motor.** The construction of this motor is similar to series motor except that its armature is short circuited on itself instead of being connected in series with stator. The operating characteristics of repulsion motor are similar to those of series motor i.e. high starting torque and high speeds at light loads. Its speed can be adjusted by shifting the brushes with the help of a special lever mechanism provided on motor. **Universal Motor**. It is a series wound motor, which operates at approximately the same speed and output on either d.c. or a.c. of approximately same voltage. The characteristics of motor is similar to those of d.c. series motors. The motor built up in fractional horse power upto about 1/4 H.P. These are suitable for sewing machines, table fans. Vacuum cleaners, portable drills, hair dryer, cooler pumps and small blowers etc. **Series Motor.** In this case also a resistance is connected in series with the armature as shown in Fig. If the current drawn by the motor is constant then flux is constant and the speed depends upon une induced e.m.f. (E). With the addition of variable resistance induced e.m.f. reduces and hence the speed. By changing the resistance, the speed below normal can be obtained. But this method is not economical.
5	Questions	1. Write application of dc motor?

		2. What is 3-phase induction motor? 3. What is single phase induction motor?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-47

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	ELECTRIC TRACTION: Explain system of traction.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: ★ What is system of traction
3	Tools used	Chalk, Blackboard
4	Lecture Description	1. The most important requirements of the driving equipment for traction service are as follows: (i) Maximum tractive effort should be exerted at starting in order that a rapid acceleration may be attained. (ii) The equipment should be capable of overloads for short periods. (iii) The wear caused on the track should be minimum. (iv) The locomotive or train unit should be self-contained and able to run on any route.
5	Questions	1. Explain system of traction?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-48

UTILIZATION OF ELECTRICAL ENERGY & TRACTION

OMR-4

TIME- 30 MINS **TOTAL MARK- 40**

1. Which of the following is an advantage of electric traction over other methods of traction ?
(a) Faster acceleration
(b) No pollution problems
(c) Better braking action
(d) All of the above
2. Which of the following is the voltage for single phase A.C. system
(a) 22 V
(b) 440 V
(c) 5 kV
(d) 15 kV
(e) None of the above
3. Long distance railways use which of the following ?
(a) 200 V D.C.
(b) 25 kV single phase A.C.
(c) 25 kV two phase A.C.
(d) 25 kV three phase A.C.
4. The speed of a locomotive is controlled by
(a) flywheel
(b) gear box
(c) applying brakes
(d) regulating steam flow to engine
5. Main traction systems used in India are, those using
(a) electric locomotives
(b) diesel engine locomotives
(c) steam engine locomotives
(d) diesel electric locomotives
(e) all of the above
6, in India diesel locomotives are manufactured at
(a) Ajmer
(b) Varanasi
(c) Bangalore
(d) Jamalpur
7. For diesel locomotives the range of horsepower is
(a) 50 to 200
(b) 500 to 1000
(c) 1500 to 2500
(d) 3000 to 5000
8. ________locomotive has the highest operational availability.
(a) Electric
(b) Diesel

(c) Steam
9. The horsepower of steam locomotives is
(a) upto 1500
(b) 1500 to 2000
(c) 2000 to 3000
(d) 3000 to 4000
10. The overall efficiency of steam locomotive is around
(a) 5 to 10 percent
(b) 15 to 20 percent
(c) 25 to 35 percent
(d) 35 to 45 percent
11. In tramways which of the following motors is used ?
(a) D.C. shunt motor
(b) D.C. series motor
(c) A.C. three phase motor
(d) AC. single phase capacitor start motor
12. In a steam locomotive electric power is provided through
(a) overhead wire
(b) battery system
(c) small turbo-generator
(d) diesel engine generator
13. Which of the following drives is suitable for mines where explosive gas exists ?
(a) Steam engine
(b) Diesel engine
(c) Battery locomotive
(d) Any of the above
14. In case of locomotives the tractive power is provided by
(a) single cylinder double acting steam engine
(b) double cylinder, single acting steam engine
(c) double cylinder, double acting steam engine
(d) single stage steam turbine
15. Overload capacity of diesel engines is usually restricted to
(a) 2 percent
(b) 10 percent
(c) 20 percent
(d) 40 percent
16. In case of steam engines the steam pressure is
(a) 1 to 4 kgf/cm2
(b) 5 to 8 kgf/cm2
(c) 10 to 15 kgf/cm2
(d) 25 to 35 kgf/cm2

17.. The steam engine provided on steam locomotives
(a) single acting condensing type
(b) single acting non-condensing type
(e) double acting condensing type
(d) double acting non-condensing type
18. Electric locomotives in India are manufactured at
(a) Jamalpur
(b) Bangalore
(c) Chittranjan
(d) Gorakhpur
19. The wheels of a train, engine as well as bogies, are slightly tapered to
(a) reduce frietion
(b) increase friction
(c) facilitate braking
(d) facilitate in taking turns
20. Automatic signalling is used for which of the following trains ?
(a) Mail and express trains
(b) Superfast trains
(c) Suburban and Urban electric trains
(d) All trains

S.K.D.A.V Govt. Polytechnic, Rourkela

Lecture Plan-49

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	System of Track electrification.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. What is system of Track electrification
3	Tools used	Chalk, Blackboard
4	Lecture Description	Electric Traction Systems The system which use electrical power for traction system i.e. for railways, trams, trolleys, etc. is called electrical traction. The track electrification refers to the type of source supply system that is used while powering the electric locomotive systems. It can be AC or DC or a composite supply. Selecting the type of electrification depends on several factors like availability of supply, type of an application area, or on the services like urban, suburban and main line

services, etc.

The three main types of electric traction systems that exist are as follows:

Direct Current (DC) electrification system

Alternating Current (AC) electrification system

Composite system.

Electrification System Types

1- DC Electrification System

The choice of selecting DC electrification system encompasses many advantages, such as space and weight considerations, rapid acceleration and braking of DC electric motors, less cost compared to AC systems, less energy consumption and so on.

In this type of system, three-phase power received from the power grids is de-escalated to low voltage and converted into DC by the rectifiers and power-electronic converters.

DC Traction System

This type of DC supply is supplied to the vehicle through two different ways:

3rd and 4 the rail system operate at low voltages (600-1200V)

Overhead rail systems use high voltages (1500-3000V)

The supply systems of DC electrification include;

300-500V supply for the special systems like battery systems.

600-1200V for urban railways like tramways and light metro trains.

1500-3000V for suburban and mainline services like light metros and heavy metro trains.

Due to high starting torque and moderate speed control, the DC series motors are extensively employed in the DC traction systems. They provide high torque at low

		speeds and low torque at high speeds. Advantages; In case of heavy trains that require frequent and rapid accelerations, DC traction motors are better choice as compared AC motors. DC train consumes less energy compared to AC unit for operating same service conditions. The equipment in DC traction system is less costly, lighter and more efficient than AC traction system. It causes no electrical interference with nearby communication lines. Disadvantages; Expensive substations are required at frequent intervals. The overhead wire or third rail must be relatively large and heavy. Voltage goes on decreasing with increase in length. 2- AC Electrification System An AC traction system has become very popular nowadays, and it is more often used in most of the traction systems due to several advantages, such as quick availability and generation of AC that can be easily stepped up or down, easy controlling of AC motors, less number of substations requirement, and the presence of light overhead catenaries that transfer low currents at high voltages, and so on. The supply systems of AC electrification include single, three phase, and composite systems. The Single phase systems consist of 11 to 15 KV supply at 16.7Hz, and 25Hz to facilitate variable speed to AC commutation motors. It uses step down transformer and

frequency converters to convert from the high voltages and fixed industrial frequency.

The Single phase 25KV at 50Hz is the most commonly used configuration for AC electrification. It is used for heavy haul systems and main line services since it doesn't require frequency conversion. This is one of the widely used types of composite systems wherein the supply is converted to DC to drive DC traction motors.

AC Traction System

Three phase system uses three phase induction motor to drive the locomotive, and it is rated at 3.3.KV, 16.7Hz. The high-voltage distribution system at 50 Hz supply is converted to this electric motor rating by transformers and frequency converters. This system employs two overhead lines, and the track rail forms another phase, but this raises many problems at crossings and junctions.

Fewer substations are required.

Lighter overhead current supply wire can be used.

Reduced weight of support structure.

Reduced capital cost of electrification.

Disadvantages;

Significant cost of electrification.

Increased maintainance cost of lines.

Overhead wires further limit clearance in tunnels.

Upgrading needs additional cost especially in case there are brigdes and tunnels.

Railway traction needs immune power with no cuts.

3- Composite System

Composite System (or multi-system) trains are used to provide continuous journeys along routes that are electrified using more than one system. One way to accomplish this is by changing locomotives at the switching stations. These stations have overhead wires that can be switched from one voltage to another. Another way is to use multi-

		system locomotives that can operate under several different voltages and current types. In Europe, it is common to use four-system locomotives. (1.5 kV DC, 3 kV DC, 15 kV 16⅔ Hz AC, 25 kV 50 Hz AC).
5	Questions	1. What is ac and dc track electrification
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-50

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	Running Characteristics of DC and AC traction motor
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. What is running Characteristics of DC traction
3	Tools used	Chalk, Blackboard
4	Lecture Description	Direct-current motors with series field windings are the oldest type of traction motors. These provide a speed-torque characteristic useful for propulsion, providing high torque at lower speeds for acceleration of the vehicle, and declining torque as speed increases. By arranging the field winding with multiple taps, the speed characteristic can be varied, allowing relatively smooth operator control of acceleration. A further measure of control is provided by using pairs of motors on a vehicle in series-parallel control; for slow operation or heavy loads, two motors can be run in series off the direct-current supply. Where higher speed is desired, these motors can be operated in parallel, making a higher voltage available at each and so allowing higher speeds. Parts of a rail system might use different voltages, with higher voltages in long runs between stations and lower voltage near stations where only slower operation is needed. A variant of the DC system is the AC series motor, also known as the universal motor, which is essentially the same device but operates on alternating current.

		Since both the armature and field current reverse at the same time, the behavior of the motor is similar to that when energized with direct current. To achieve better operating conditions, AC railways are often supplied with current at a lower frequency than the commercial supply used for general lighting and power; special traction current power stations are used, or rotary converters used to convert 50 or 60 Hz commercial power to the 25 Hz or 16+2/3 Hz frequency used for AC traction motors. The AC system allows efficient distribution of power down the length of a rail line, and also permits speed control with switchgear on the vehicle. AC induction motors and synchronous motors are simple and low maintenance, but are awkward to apply for traction motors because of their fixed speed characteristic. An AC induction motor generates useful amounts of power only over a narrow speed range determined by its construction and the frequency of the AC power supply. The advent of power semiconductors has made it possible to fit a variable frequency drive on a locomotive; this allows a wide range of speeds, AC power transmission, and rugged induction motors without wearing parts like brushes and commutators.[1]
5	Questions	. explain running characteristics of ac traction motor?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-51

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	Explain control of motor: 1. Tapped field control. 2. Rheostatic control. 3. Series parallel control.

		4. Multi-unit control. 5. Metadyne control
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1. What is series parallel control 2. What is Metadyne control
3	Tools used	Chalk, Blackboard
4	Lecture Description	Tapped Field Control This is another method of increasing the speed by reducing the flux and it is done by lowering number of turns of field winding through which current flows. In this method a number of tapping from field winding are brought outside. This method is employed in electric traction. Rheostatic control: Rheostatic control is the sim•plest but the least efficient method of speed control of 3-phase in•duction motors. This method of speed control is employed in light locomotives and motor coaches where a single economical speed is sufficient and energy consump•tion is of no importance. Pole-changing control is the simplest of the multi-speed control methods. This method has the advantage of simplicity, good speed regulation for each setting, high operation efficiency, and moderate first cost and maintenance. The choice of the number of poles on a pole-changing winding is in the ratio of 2 : 1, 3 : 2, 4 : 3. Cascade control is employed for the two-speed goods locomotives (24 kmph and 50 kmph). In combination of cascading and pole-changing control method, both of the motors employed are wound for the same number of poles and windings so arranged that the number of poles can be changed. With equal number of poles on both the machines one speed is obtained and then more speeds are made available by changing the number of poles. This method has been employed for the four-speed passenger locomotives (37, 51, 74 and 102 kmph) in some types of locomotives. **Series parallel control**

		This series parallel method is generally employed with the resistance control. For example two series motor are mechanically coupled to drive a vehicle. At cessation, through a starting resistance the motors are connected in series. The series motors are taking place up in series respectively and the starting resistance is gradually cut out step by step in rise the speed. When the total resistance is cutout step by step the voltage supplied to each one of the motor is about one-half of the line voltage and the speed is also increased about one-half times. Further to increase the motor speed, then each motors are to be connected in parallel and the resistance to be connected in series at the same time. The starting resistance is again gradually reduced till full speed of the motor is achieved. At that moment field control is attained. **Multi- Unitl control** Multiple-unit train control, sometimes abbreviated to multiple-unit or MU, is a method of simultaneously controlling all the traction equipment in a train from a single location—whether it is a multiple unit comprising a number of self-powered passenger cars or a set of locomotives—with only a control signal transmitted to each unit. This contrasts with arrangements where electric motors in different units are connected directly to the power supply switched by a single control mechanism, thus requiring the full traction power to be transmitted through the train. A set of vehicles under multiple unit control is referred to as a consist in the United States. Metadyne control: Multiple-unit train control, sometimes abbreviated to multiple-unit or MU, is a method of simultaneously controlling all the traction equipment in a train from a single location—whether it is a multiple unit comprising a number of self-powered passenger cars or a set of locomotives—with only a control signal transmitted to each unit. This contrasts with arrangements where electric motors in different units are connected directly to the power supply switched by a single control mechanism, thus requiring the full traction power to be transmitted through the train. A set of vehicles under multiple unit control is referred to as a consist in the United States.
5	Questions	1. What is multi unit control ? 2. Explain rheostatic control? 3. What is metadyne control

6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

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Lecture Plan-52

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	Explain Braking of the following types: 1. Regenerative Brakin 2. Braking with 1-phase series motor. 3. Magnetic Braking.
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1.
3	Tools used	Chalk, Blackboard
4	Lecture Description	**Regenerative Brakin** Regenerative braking. In the first two methods, the mechani cal energy of the revolving parts was converted into electrical energy which was further dissipated in the external resistance. In case of plugging, even energy was drawn during the braking period also and was wasted. (In case of regenerative braking, the machine is made to work in a way that it generates energy and feeds it back to the supply system. As shown in Fig. 9.41 the shunt machine draws power from supply if applied voltage 'V' is greater than induced voltages E. If induced voltage E, becomes greater than the applied voltage, the machine instead of A taking power from the supply (motoring action) sends back power to supply (generating action). This can be achieved if, (i) Supply voltage T is reduced. (ii) The field excitation is increased so that E, becomes greater than applied

voltage V.

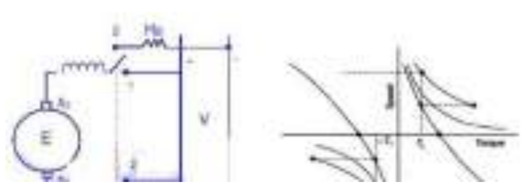
(iii) The speed of the motor becomes greater than the no-load. speed of the motor, in that case again the value of Es becomes greater than the applied voltage V.

Thus whenever the value of E, becomes greater than V the direction of current in the armature gets reversed, the direction of field remaining same, the torque produced is in the opposite direc tion. Thus braking effect is produced. When the torque is reversed, the motor tends to stop thus reducing the speed and thus reducing Es. In order that braking torque is still applied, the field excitation is increased so that Es still remains greater than applied voltage. Thus it is not possible to obtain regenerative braking down to very low speeds as E, falls tremendously down and it cannot be increased by the increase in excitation because too much increase in field cannot be obtained.

Thus regenerative braking cannot stop the motor. It can only be applied to overhauling loads which tend to drive the motor at speeds greater than the no-load speed of the motor. This can take place in a crane motor due to action of lowering load or in a motor of an electric locomotive moving down-gradient.

Braking with 1-phase series motor

In this method of breaking, the motor is disconnected from the supply, the field connections are reversed and motor is connected in series with a variable resistance R as shown in



		The field connections are reversed to make sure that,the current through the field winding flows in the same direction as before (i.e., from A to B)in order to assist for residual magnetism. • In practice, the variable resistance used for starting purpose is itself used for braking purposes. • The speed-torque characteristics of DC series motor during rheostatic braking is shown in the following figure. explanations are similar to rheostatic braking method applied to DC shunt motor. **Magnetic Braking** In this case the excitation of the armature is disconnected from the supply but the excitation remains on. When the armature rotates in the fixed field, there is reversal of flux in the armature and the iron losses are fed from the kinetic energy of the rotating components and the machine retards. This method can be adopted for shunt, compound „nd synchronous motors. In case of series motors the field cannot stand the full rated voltage, so separate battery has to provided for excitation during braking.
5	Questions	1. What is regenerative braking? 2. Explain magnetic Braking?
6	Reference	https://thebookee.net/ut/utilization-of-electric-power-and-electric-traction-by-g-c-garg Utilisation of electric power and electric traction by G.C. Garg

S.K.D.A.V Govt. Polytechnic, Rourkela

Lecture Plan-53

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1.
3	Tools used	Chalk, Blackboard
4	Lecture Description	
5	Questions	
6	Reference	

S.K.D.A.V Govt. Polytechnic, Rourkela

Lecture Plan-54

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	
2	Intended Learning Outcome	On successful completion of the Lecture students are able to: 1.
3	Tools used	Chalk, Blackboard and PPT
4	Lecture Description	
5	Questions	
6	Reference	

S.K.D.A.V Govt. Polytechnic, Rourkela

Lecture Plan-55

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	
2	Intended Learning Outcome	On successful completion of the Lecture students are able to:
3	Tools used	Chalk, Blackboard
4	Lecture Description	
5	Questions	
6	Reference	

S.K.D.A.V Govt. Polytechnic, Rourkela

Lecture Plan-56

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	OMR -5
2	Intended Learning Outcome	On successful completion of the Lecture students are able to:
3	Tools used	Chalk, Blackboard
4	Lecture Description	

5	Questions	
6	Reference	

S.K.D.A.V Govt. Polytechnic, Rourkela

Lecture Plan-57

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	
2	Intended Learning Outcome	On successful completion of the Lecture students are able to:
3	Tools used	Chalk, Blackboard
4	Lecture Description	
5	Questions	
6	Reference	

S.K.D.A.V Govt. Polytechnic, Rourkela

Lecture Plan-58

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	
2	Intended Learning Outcome	On successful completion of the Lecture students are able to:
3	Tools used	Chalk, Blackboard
4	Lecture Description	
5	Questions	
6	Reference	

S.K.D.A.V Govt. Polytechnic, Rourkela

Lecture Plan-59

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	
2	Intended Learning Outcome	On successful completion of the Lecture students are able to:
3	Tools used	Chalk, Blackboard
4	Lecture Description	
5	Questions	
6	Reference	

S.K.D.A.V Govt. Polytechnic, Rourkela

Lecture Plan-60

Subject -UTILIZATION OF ELECTRICAL ENERGY & TRACTION Semester- 5th		
1	Topic	
2	Intended Learning Outcome	On successful completion of the Lecture students are able to:
3	Tools used	Chalk, Blackboard
4	Lecture Description	
5	Questions	
6	Reference	

S.K.D.A.V Govt. Polytechnic, Rourkela

Lecture Plan-61

UNIT DISCUSSION- III, IV

1. How do you describe control gear?
2. What are the advantages of automatic control over the manual control?
3. Describe briefly the starting and speed control of a.c. and d.c. motors.
4. Describe the Ward-Leonard method of speed control and illustrates its applications.?

5. Describe the principle, construction and working of the amplidyne illustrating its applications.?
6. State the advantages and disadvantages of electrical drive over mechanical drive,?
7. What type of drive is being used in modern day industry. A group drive' or 'an individual drive" Discuss the advantages of one over the other.
8. Discuss briefly the various factors which are required to be considered for selecting a motor for a particular industry.?
9. Sketch and explain the performance of D.C. series motor and hence explain the applications for which it is most suitable.?
10. Draw and explain the speed current and torque current characteristics in respect of D.C. series motor and D.C. shunt motor.
11. Write down the various methods of controlling the speed of D.C. motors.?
12. Explain the method of rheostatic control for controlling the speed of D.C. shunt motor?.
13. Explain the tap changing method of speed control of D.C. series motor?
14. Explain how the speed of D.C. shunt motor can be brought above or below the rated speed.?
15. State the various methods of speed control of three-phase induction motor. Explain any one method in detail.?
16. What is the necessity of starter in induction motor?
17. Explain 3-phase induction motor?
18. Explain 3-phase slipring induction motor?
19. What is single phase motor?
20. Compare electric braking system to that of mechanical braking system ?
21. What are the various methods of braking dc motor ?.
22. Explain the process of ' plugging' as applied to dc series motor ?
23. Distinguish between rheostatic and regeneration bracking as applied to electric traction. Under what circumstances the mechanical or regenerative braking applied?
24. Explain how induction motor is stopped by plugging?
25. Explain how induction motor is stopped by rheostatic braking ?
26. State giving reasons whether you would choose plugging or rheo static braking for quickly braking to rest :

    (i) 3 phase induction motor, (ii) Synchronous motor.
27. Describe the following features pertaining to the relation for motor:
28. (a) Types of enclosures (b) Bearings (c) Transmission of drive and noise.
29. Why D.C. series motor is preferred over other types of D.C. motors for use in electric tractions?

30. What do you understand by 'series' and 'shunt' speed-torque char acteristics of electric motors used in industrial drives? Name (at least one of each type) d.c. motors and a.c. motors with such characteristics.
31. How does a 3-phase synchronous motor differ from a 3 o induction motor as far as the starting and running features are concerned? With the help of suitable diagrams, explain the different methods used in industrial drives to start a 3-phase synchronous motor.
32. Discuss the merits and demerits of using :

(a) Star-delta starter and (b) Auto-transformer starter with a 3-phase cage induction motor.

33. Describe in brief speed control of induction motor by pole changing.
34. A motor is to be selected for driving a load having large torque of short duration followed by a long no-load period. A flywheel of suitable inertia is already mounted on the load shaft. Out of the three types of D.C. motors (series, shunt and compound) which will you recommend for this application and why?
35. State giving reasons what type of motor would be suitable for the following drives

    1. Propeller shaft of a ship
    2. Rotary compressors.
    3. Rotary printing press
    4. Spinning machine in textile mills

    Would you suggest any load equaliser for the above drives? Justify.
36. What do you mean by load equalisation? How is it affected?
37. Why the D.C. series motor is used for traction application?
38. Why is a series D.C. motor not to be run on no load?
39. Why electric drive is preferred over mechanical drive?
40. Why individual drive is preferred over group drive now-a-days ?
41. State the main features for an ideal traction system?
42. State the advantages of electric drive in case of traction system.
43. What are the advantages of electric traction over other forms of traction with special reference to the suburban traffic? What are technical aspects of railway electrification?
44. State the merits and/or demerits of electric traction with reference to Elecrtic Power Crisis in India.
45. Discuss the various systems of track electrifcation:
46. What is dielectric traction?
47. Explain traction system in India
48. What is battery locomotive?

49. Explain dc versus ac traction
50. Write single phase traction at 50 Hz
51. Explain single phase low frequency ac system of electrification?
52. What do you understand by speed-time curves? What is its use practice?
53. Define crest speed, average speed and schedule speed and discuss the factors which effect the schedule speed of a train.
54. Explain clearly "Free running", "Coasting' and 'Braking' with refer ence to electric traction system.
55. Draw the speed time curve for an electric train accelerating uniformly to a speed Vm. the power is cut off and after a time t when it attains Vm, coasting is allowed to continue up to the end of the run, the braking being neglected. Drive a suitable equation for Vm.
56. Explain how an actual speed-time curve for an electric train service can be replaced by a curve having simple geometric shape deduce, from first principles, the relationship between acceleration, maximum speed, running time and distance between stops assuming a simplified speed-time curve.
57. Derive an expression for the tractive effort.
58. Explain the terms, "Dead weight", "Effective weight" and "Adhesive weight" in a locomotive.
59. For a typical trapezoidal speed-time curve, show how you would calculate the energy consumed during two consecutive stops.Use either graphical or semi-graphical method.
60. Explain the terms "Specific energy output" and specific energy consumption.
61. Explain the terms Adhesive weigh and co-efficient of adhesion.
62. Explain briefly how regenerative braking differs from other methods of braking traction motors. With neat and explanatory notes, demonstrate the usefulness of this method of braking with (a) D.C. series motors (b) A.C. motors in traction.
63. Explain electric braking?
64. What is dynamic Braking?
65. Explain regenerative braking?
66. Describe the main difficulties encountered in regenerative braking of D.C. series motor and describe any one scheme for over coming these difficulties.
67. Describe the power circuit of the driving motor for 25 kW, 1-phase,50 Hz a.c. traction.
68. Write Different methods of electric braking in a traction system.
69. What do you understand by tractive effort? Derive an expression for the same.
70. Discuss in details why series motors are ideal for D.C. or A.C. traction.
71. What is catenary?
72. What is pantography ?
73. Explain circuit breakers?

74. Explain smoothing reactor?
75. Define on load tap changer?
76. What is transformer?
77. What is rectifier?
78. Explain dc motor used for traction?

GANDHI INSTITUTE OF EXCELLENT TECHNOCRATS
BHUBANESWAR

ELECTRICAL ENGINEERING

LECTURE NOTES

Electrical Measurement & Instrumentation

Subject code – Th.3

Semester-4th

Prepared By:-

CONTENTS

CHAPTER-1

MEASURING INSTRUMENTS

1. **ACCURACY-** The closeness with which an instrument reading approaches the true value of the quantity being measured is called accuracy. Accuracy is determined as the maximum amount by which the result differs from the true value.

2. **PRECISION-** The term precise means means clearly or sharply defined. Precision is a measure of reproducibility of measurement.

3. **ERRORS-** The deviation or change of the value obtained from measurement from the desired standard value.

Mathematically,

Error = Obtained Reading/Value – Standard Reference Value.

There are three types of error. They are as follows:-

GROSS ERRORS-This are the error due to humans mistakes such as careless reading mistakes in recoding observation incorrect application of an instrument.

SYSTEMATIC ERROR-A constant uniform deviation of an instrument is as systematic error. There are two types of systematic error.

a) STATIC ERROR

Thestatic error of a measuring instrument is the numerical different between the true value of a quantity and its value as obtained by measurement.

b) DYNAMIC ERROR-

1. It is the different between true value of a quantity changing with and value indicated by the instrument.
2. The Dynamic Errors are caused by the instrument not responding fast enoughto follow the changes in the measured value.

RANDOM ERROR-The cause of such error is unknown or not determined in the ordinary process of making measurement.

TYPES OF STATIC ERROR

1. INSTRUMENTAL ERROR- Instrumental error are errors inherent in mastering

instrument because of the mechanical construction friction is bearing in various moving component. It can be avoided by

a. Selecting a suitable instrument for the particular measurement.

b. Applying correction factor after determining the amount of instrumental error.

2. ENVIROMENTAL ERROR –Environmental error are due to conditions external tothe measuring device including condition al in the area surrounding the instrument such as effect of change in temperature , humidity or electrostatic field it can be avoided

a. Providing air conditioning.

b. Use of magnetic shields.

3. OBSERVATIONAL ERROR- The errors introduced by the observer. These errors are caused by habits of the observers like tilting his/her head too much while reading a "Needle – Scale Reading".

Measuring instruments are classified according to both the quantity measured by the instrument and the principle of operation.

There are three general principles of operation:

- electromagnetic, which utilizes the magnetic effects of electric currents;
- electrostatic, which utilizes the forces between electrically-charged conductors;
- Electro-thermic, which utilizes the heating effect

Electric measuring instruments and meters are used to indicate directly the value of current,voltage, power or energy.An electromechanical meter (input is as an electrical signal results mechanical force or torqueas an output) that can be connected with additional suitable components in order to act as anammeters and a voltmeter.The most common analogue instrument or meter is the permanent magnet moving coilinstrument and it is used for measuring a dc current or voltage of an electric circuit.

4. SENSITIVITY- Sensitivity can be defined as a ratio of a change output to the change input at steady state condition.

5. RESOLUTION- Resolutions the least increment value of input or output that can be detected, caused or otherwise discriminated by the measuring device.

6. TOLETANCE- Tolerance refers to the total allowable error within an item. This is typically represented as a +/- value off of a nominal specification. Products can become deformed due to changes in temperature and humidity, which lead to material expansion and contraction, or due to improper feedback from a process control device. As such, it's necessary to take errors into consideration with regard to design values in the manufacturing

and inspection processes.

CLASSIFICATION OF MEASURING INSTRUMENT-

The instrument may be classified as

1. Absolute and secondary Instrument
2. Analog and digital Instrument
3. Mechanical, Electrical and Electronics Instruments.
4. Manual and automatic instruments
5. Self contained and remote indicating instruments
6. Self operated and power-operated instrument
7. Deflection and Null output instrument.

1. Absolute instrument & Secondary instruments:

Absolute instrument measures the process variable directly from the process without the use of conversion. Such instruments do not require comparison with any other standard. The tangent galvanometer is an example for the absolute instrument. These instruments are used as standards in labs and institution.

Secondary instrument: These instruments are so constructed that the deflection of such instruments gives the magnitude of the electrical quantity to be measured directly. These instruments required to calibrated with respect to the standard instrument. These instruments are usually used in practice.

Secondary instruments are classified as:

1. Indicating instrument: Those instruments that measure and indicates the magnitude of the electricity. The indications are given by a pointer moving over a calibrated scale. Ordinary ammeters, voltmeters, wattmeters, frequency meters, power factor meters, etc., fall into this category.
2. Integrating instrument: Integrating instruments are those which measure the total amount of either quantity of electricity (ampere-hours) or electrical energy supplied over a period of time. The ampere-hour meters and energy meters fall in this class.
3. Recording instrument: These instruments record continuously the variation of the magnitude of the electric quantity for a definite period of time. Such instruments are generally used in powerhouses where the current, voltage, power, etc., are to be maintained within a certain acceptable limit.

2. Analog and digital Instrument

Analog instrument: The signals of an analog unit vary in a continuous fashion and can take on an infinite number of values in a given range. Fuel gauge, ammeter and voltmeters, wristwatch, speedometer fall in this category.

Digital Instruments: Signals that vary in discrete steps and that take a finite number of different values in a given range are digital signals and the corresponding instruments are of digital type. Digital instruments have some advantages over analog meters, in that they have high accuracy and high speed of operation. Digital multimeter is an example for the digital instrument.

3. Mechanical, Electrical and Electronics Instruments.

Mechanical instrument: Mechanical instruments are very reliable for static and stable conditions. As they use mechanical parts these instruments cannot faithfully follow the rapid changes which are involved in dynamic instruments. But they are cheaper in cost and durable.

Electrical Instruments: When the instrument pointer deflection is caused by the action of some electrical methods then it is called an electrical instrument. The time of operation of an electrical instrument is more rapid than that of a mechanical instrument. This mechanical movement has some inertia due to which the frequency response of these instruments is poor.

Electronic Instruments: Electronic instruments use semiconductor devices. They are very fast in response. In electronic devices, since the only movement involved is that of electrons, the response time is extremely small owing to very small inertia of the electrons. With the use of electronic devices, a very weak signal can be detected by using pre-amplifiers and amplifiers.

4. Manual and automatic instruments

In case of ***Manual instruments*** services of an operator are required. Example: Measurement of temperature by a resistance thermometer incorporating a Wheatstone bridge in its circuit.

In an ***Automatic Instrument*** an operator is not required. Example: Measurement of temperature by mercury-in-glass thermometer.

5. Self contained and remote indicating instruments

A ***Self contained instrument*** has all its different elements in one physical assembly.

In a ***Remote Indicating Instrument*** the primary sensing element may be located at an adequate long distance from the secondary indicating element. Such type of instrument are finding wide use in the modern instrumentation technology.

6. Self operated and power-operated instrument

Self-operated instruments don't need any outside power for its working. The output energy is supplied wholly or almost wholly by the input measurand. Dial indicating type instruments belong to this category.

Power operated instrument need external power for its working. External power can electric current, hydraulic or pneumatic energy. In such cases, the input signal supplies only an insignificant portion of the output power.

7. Deflection and Null output instrument.

In a ***deflection-type instrument***, the deflection of the instrument indicates the measurement of the unknown quantity. The measurand quantity produces some physical effect which deflects or produces a mechanical displacement in the moving system of the instrument.

An opposite effect is built in the instrument which opposes the deflection or the mechanical displacement of the moving system. The balance is achieved when opposing effect equals the actuating cause producing the deflection or the mechanical displacement. Permanent Magnet Moving Coil (PMMC), Moving Iron (MI), etc., type instruments are examples of this category.

In ***Null type instruments***, a zero or null indication leads to the determination of the magnitude of the measurand quantity. The null condition depends upon some other known conditions. These are more accurate and highly sensitive as compared to deflection-type instruments. A dc potentiometer is a null- type instrument.

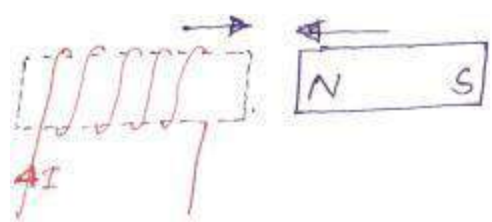
TYPES OF FORCES/TORQUES ACTING IN MEASURING INSTRUMENTS

1. DEFLECTING TORQUE/FORCE:

The defection of any instrument is determined by the combined effect of thedeflecting torque/force, control torque/force and damping torque/force.The value of deflecting torque must depend on the electrical signal to be measured.This torque/force causes the instrument movement to rotate from its zero position.

MAGNITUDE EFFECT

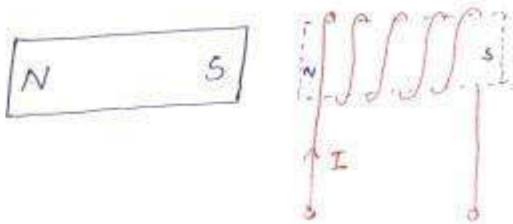
When a current passes through the coil, it produces a imaginary bar magnet. When a soft-iron piece is brought near this coil it is magnetized. Depending upon the current direction the poles are produced in such a way that there will be a force of attraction between the coil and the soft iron piece. This principle is used in moving iron attraction type instrument.



If two soft iron pieces are place near a current carrying coil there will be a force of repulsion between the two soft iron pieces. This principle is utilized in the moving iron repulsion type instrument.

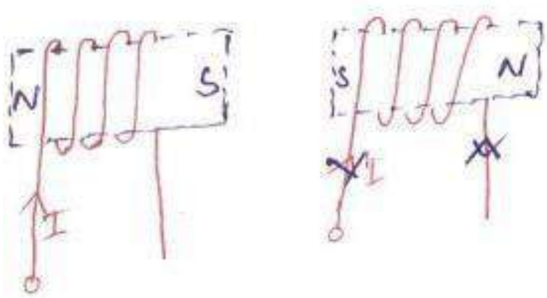
FORCE BETWEEN A PERMANENT MAGNET AND A CURRENT CARRYING COIL

When a current carrying coil is placed under the influence of magnetic field produced by a permanent magnet and a force is produced between them. This principle is utilized in the moving coil type instrument.



FORCE BETWEEN TWO CURRENT CARRYING COIL

When two current carrying coils are placed closer to each other there will be a force of repulsion between them. If one coil is movable and other is fixed, the movable coil will move away from the fixed one. This principle is utilized in electrodynamometer type instrument.



2. CONTROLLING TORQUE/FORCE:

This torque/force must act in the opposite sense to the deflecting torque/force, and themovement will take up an equilibrium or definite position when the deflecting andcontrolling torque are equal in magnitude.The Spiral springs or gravity usually provides the controlling torque.
When the external signal to be measured by the instrument is removed, the pointer should return back to the zero position. This is possibly due to the controlling force and the pointer will be indicating a steady value when the deflecting torque is equal to controlling torque.

$$Td = Tc$$

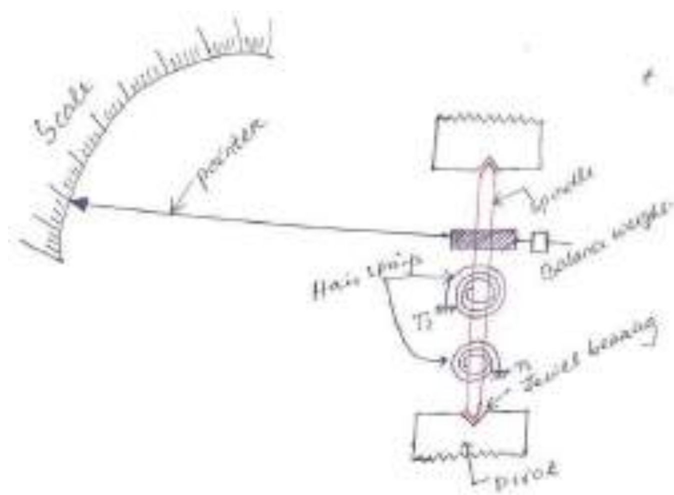
SPRING CONTROL

Two springs are attached on either end of spindle. The spindle is placed in jewelled bearing, so that the frictional force between the pivot and spindle will be minimum. Two springs are provided in opposite direction to compensate the temperature error. The spring is made of phosphorous bronze. When a current is supply, the pointer deflects due to rotation of the spindle. While spindle is rotate, the spring attached with the spindle will oppose the movements of the pointer. The torque produced by the spring is directly proportional to the pointer deflectionθ.

$$T_C a \theta$$

The deflecting torque produced T_d proportional to 'I'. When $T_C = T_d$, the pointer will come to a steady position. Therefore

$\theta \propto I$



3. DAMPING TORQUE/FORCE:

A damping force is required to act in a direction opposite to the movement of the moving system.This brings the moving system to rest at the deflected position reasonably quickly withoutany oscillation or very small oscillation.

To damp out the oscillation is quickly, a damping force is necessary.
This force is produced by different systems.

(a) Air friction damping
(b) Fluid friction damping
(c) Eddy current damping

CHAPTER- II

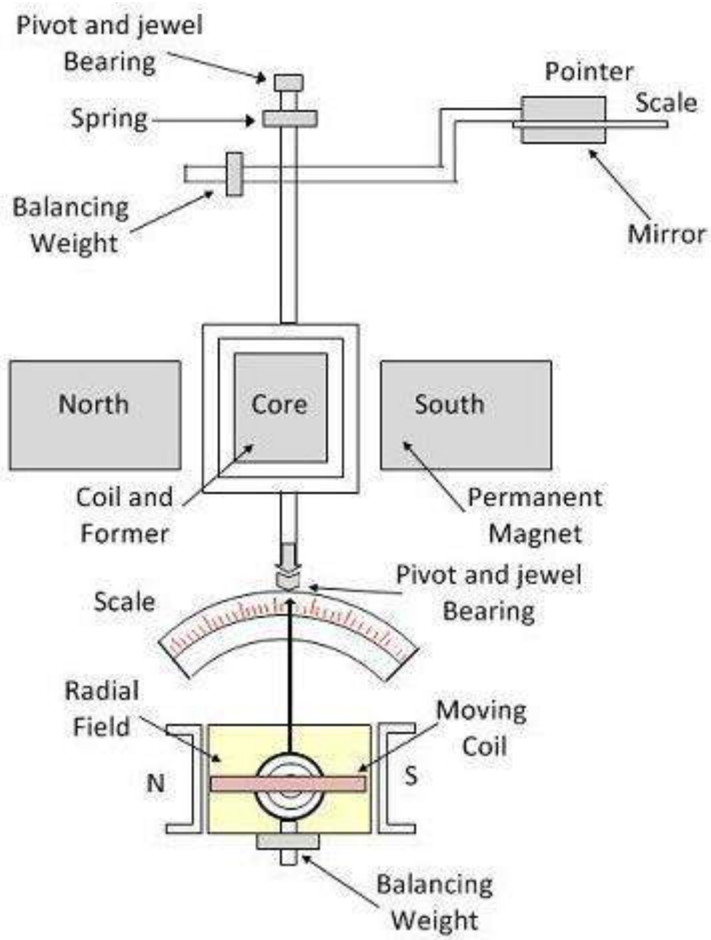
ANALOG AMMETERS & VOLTMETERS

PERMANENT MAGNET MOVING COIL (PMMC) INSTRUMENT

One of the most accurate type of instrument used for D.C. measurements is PMMC instrument.

CONSTRUCTION:

The moving coil and permanent magnet are the main part of the PMMC instrument. The parts of the PMMC instruments are explained below in details.



Permanent Magnet Moving Coil Instrument

Moving Coil – The coil is the current carrying part of the instruments which is freely moved between the stationary field of the permanent magnet. The current passes through the coil deflects it due to which the magnitude of the current or voltage is determined. The coil is mounted on the rectangular former which is made up of aluminium. The former increases the radial and uniform magnetic field between the air gaps of the poles. The coil is wound with the silk cover copper wire between the poles of a magnet.

The coil is mounted on the rectangular former which is made up of aluminium. The former increases the radial and uniform magnetic field between the air gaps of the poles. The coil is wound with the silk cover copper wire between the poles of a magnet.

Magnet System – The PMMC instrument using the permanent magnet for creating the stationary magnets. The Alcomax and Alnico material are used for creating the permanent magnet because this magnet has the high coercive force (The coercive force changes the magnetisation property of the magnet). Also, the magnet has high field intensities.

Control – In PMMC instrument the controlling torque is because of the springs. The springs are made up of phosphorous bronze and placed between the two jewel bearings. The spring also provides the path to the lead current to flow in and out of the moving coil. The controlling torque is mainly because of the suspension of the ribbon.

Damping – The damping torque is used for keeping the movement of the coil in rest. This damping torque is induced because of the movement of the aluminium core which is moving between the poles of the permanent magnet.

Pointer & Scale – The pointer is linked with the moving coil. The pointer notices the deflection of the coil, and the magnitude of their deviation is shown on the scale. The pointer is made of the lightweight material, and hence it is easily deflected with the movement of the coil. Sometimes the parallax error occurs in the instrument which is easily reduced by correctly aligning the blade of the pointer.

PRINCIPLE OF OPERATION

When D.C. supply is given to the moving coil, D.C. current flows through it. When the current carrying coil is kept in the magnetic field, it experiences a force. This force produces a torque and the former rotates. The pointer is attached with the spindle. When the former rotates, the pointer moves over the calibrated scale. When the polarity is reversed a torque is produced in the opposite direction. The mechanical stopper does not allow the deflection in the opposite direction. Therefore the polarity should be maintained with PMMC instrument. If A.C. is supplied, a reversing torque is produced. This cannot produce a continuous deflection. Therefore this instrument cannot be used in A.C.

TORQUE DEVELOPED BY PMMC

The deflecting torque induces because of the movement of the coil. The deflecting torque is expressed by the equation shown below.

$$T_d = NBLdI \ldots\ldots equ(1)$$

Where, I- Current through the coil,
N – Number of turns of coil
B – flux density in the air gap
L, d – the vertical and horizontal length of the side

$$G = NBLd \ldots\ldots equ(2)$$

The spring provides the restoring torque to the moving coil which is expressed as

$$T_C = K\theta \ldots\ldots equ(3)$$

Where K = Spring constant.

For final deflection, $T_c = T_d$

By substituting the value of equation (1) and (3) we get,

$$K\theta = GI$$

$$\theta = \frac{GI}{K} \ldots . equ(4)$$

$$I = \frac{K}{G}\theta \ldots\ldots equ(5)$$

The above equation shows that the deflection torque is directly proportional to the current passing through the coil.

ERRORS IN PMMC INSTRUMENT:

In PMMC instruments the error occurs because of the ageing and the temperature effects of the instruments. The magnet, spring and the moving coil are the main parts of the instruments which cause the error. The different types of errors of the instrument are explained below in details.

1. Magnet – The heat and vibration reduce the lifespan of the permanent magnet. This treatment also reduced the magnetism of the magnet. The magnetism is the property of the attraction or repulsion of the magnet. The weakness of the magnet decreases the deflection of the coil.

2. Springs – The weakness of the spring increases the deflection of moving coil between the permanent magnet. So, even for the small value of current, the coil show large deflection. The spring gets weakened because of the effect of the temperature. One degree rise in temperature reduces the 0.004 percent life of the spring.

3. Moving Coil – The error exists in the coil when their range is extended from the given limit by the use of the shunt. The error occurs because of the change of the coil resistance on the shunt resistance. This happens because the coil is made up of copper wire which has high shunt resistance and the shunt wire made up of Magnin has low resistance.

To overcome from this error, the swamping resistance is placed in series with the moving coil. The resistor which has low-temperature coefficient is known as the swamping resistance. The swamping resistance reduces the effect of temperature on the moving coil.

Advantages of PMMC Instruments

The following are the advantages of the PMMC Instruments.

1. The scale of the PMMC instruments is correctly divided.
2. The power consumption of the devices is very less.
3. The PMMC instruments have high accuracy because of the high torque weight ratio.
4. The single device measures the different range of voltage and current. This can be done by the use of multipliers and shunts.
5. The PMMC instruments use shelf shielding magnet which is useful for the aerospace applications.

Disadvantages of PMMC Instruments

The following are the disadvantages of the PMMC instruments.

1. The PMMC instruments are only used for the direct current. The alternating current varies with the time. The rapid variation of the current varies the torque of the coil. But the pointer can not follow the fast reversal and the deflection of the torque. Thus, it cannot use for AC.
2. The cost of the PPMC instruments is much higher as compared to the moving coil instruments.

The moving coil itself provides the electromagnetic damping. The electromagnetic damping opposes the motion of the coil which is because of the reaction of the eddy current and the magnetic field.

RANGES OF PMMC INSTRUMENT:

DC ammeter:

1. Without shunt- 0/5 micro amperes upto0/30 microamperes
2. With internal Shunt- upto 0/2000 amperes
3. With external Shunt- upto 0/5000 amperes

DC Voltmeters:
1. Without series resistance- 0/100 milli-volts
2. With series resistance- upto 0/5000 amperes

MOVING IRON (MI) INSTRUMENTS

One of the most accurate instruments used for both AC and DC measurement is moving iron instrument. There are two types of moving iron instrument.

- Attraction type
- Repulsion type

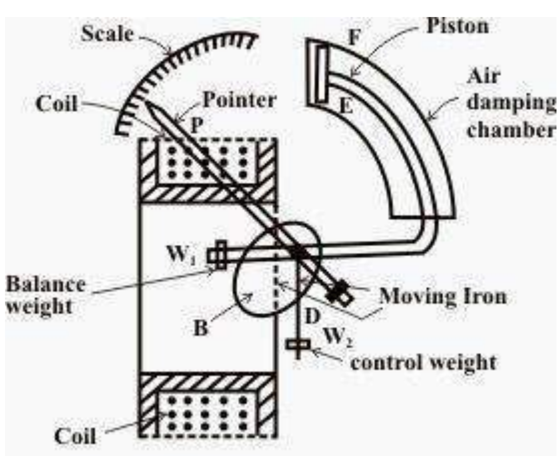
ATTRACTION TYPE M.I. INSTRUMENT

CONSTRUCTION:

It consists of a flat cylindrical coil. The moving iron is a flat disc or oval shaped disc, pivoted on a spindle. A pointer is attached to the spindle which moves on a calibrated scale. The controlling force is obtained by gravity control system. The damping force is provided by air friction with the help of light aluminium piston attached to the moving system

PRINCIPLE OF OPERATION

The current to be measured is passed through the fixed coil. As the current is low through the fixed coil, a magnetic field is produced. By magnetic induction the moving iron gets magnetized. The north pole of moving coil is attracted by the south pole of fixed coil. Thus the deflecting force is produced due to force of attraction. Since the moving iron is attached with the spindle, the spindle rotates and the pointer moves over the calibrated scale. But the force of attraction depends on the current flowing through the coil.



Attraction type

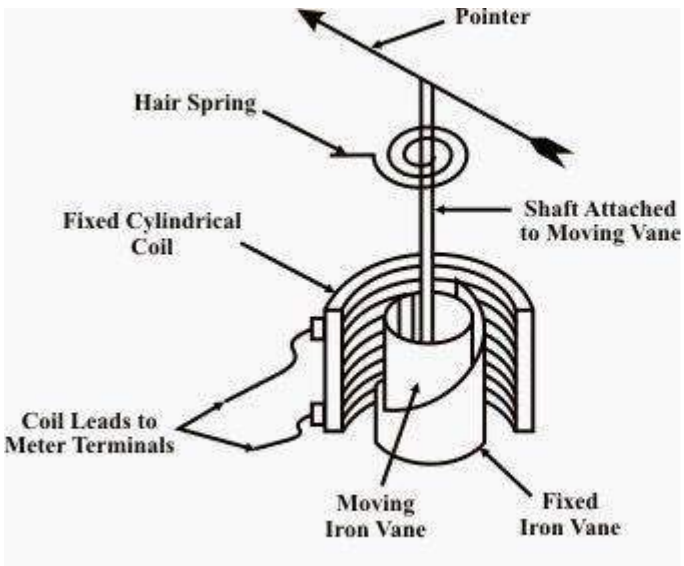
REPULSION TYPE MOVING IRON INSTRUMENT

CONSTRUCTION**:**

The repulsion type instrument has a hollow fixed iron attached to it . The moving iron is connected to the spindle. The pointer is also attached to the spindle in supported with jeweled bearing.

PRINCIPLE OF OPERATION**:**

When the current flows through the coil, a magnetic field is produced by it. So both fixed iron and moving iron are magnetized with the same polarity, since they are kept in the same magnetic field. Similar poles of fixed and moving iron get repelled. Thus the deflecting torque is produced due to magnetic repulsion. Since moving iron is attached to spindle, the spindle will move. So that pointer moves over the calibrated scale. Damping: Air friction damping is used to reduce the oscillation. Control: Spring control is used.



Repulsion type.

Ranges of Ammeter and Voltmeter

1. For a given moving-iron instrument the ampere-turns necessary to produce full-scale deflection are constant.
2. One can alter the range of ammeters by providing a shunt coil with the moving coil.
3. Voltmeter range may be altered connecting a resistance in series with the coil. Hence the same coil winding specification may be employed for a number of ranges.

Advantages

1. The instruments are suitable for use in AC and DC circuits.
2. The instruments are robust, owing to the simple construction of the moving parts.
3. The stationary parts of the instruments are also simple.
4. Instrument is low cost compared to moving coil instrument.
5. Torque/weight ratio is high, thus less frictional error.

Disadvantages

1. Scale not uniform.
2. For low voltage range, the power consumption is higher.
3. The errors are caused due to hysteresis in the iron of the operating system and due to stray magnetic field.
4. In case of AC measurements, change in frequency causes serious error.
5. With the increase in temperature the stiffness of the spring decreases.

Errors

- Error due to variation in temperature.
- Error due to friction is quite small as torque-weight ratio is high in moving coil instruments.
- Stray fields cause relatively low values of magnetizing force produced by the coil. Efficient magnetic screening is essential to reduce this effect.
- Error due to variation of frequency causes change of reactance of the coil and also changes the eddy currents induced in neighbouring metal.
- Deflecting torque is not exactly proportional to the square of the current due to non-linear characteristics of iron material.

DYNAMOMETER (OR) ELECTROMAGNETIC MOVING COIL INSTRUMENT

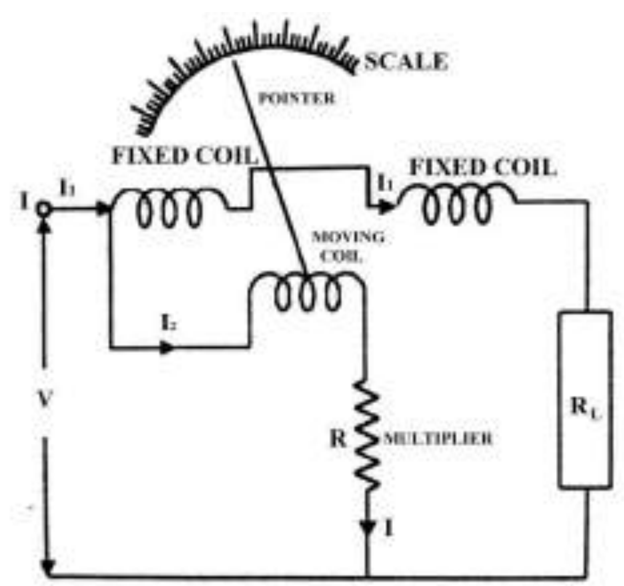
Dynamometer type measuring instruments are similar to PMMC instrument. Except that the permanent magnetic field coil is replaced by a coil which carries the current to be measured. They have precision grade accuracy both for ac and dc measurements

CONSTRUCTION:

A fixed coil is divided in to two equal half. The moving coil is placed between the two half of the fixed coil. Both the fixed and moving coils are air cored. So that the hysteresis effect will be zero. The pointer is attached with the spindle. In a non metallic former the moving coil is wounded.

Control: Spring control is used.

Damping: Air friction damping is used.



PRINCIPLE OF OPERATION:

When the current flows through the fixed coil, it produced a magnetic field, whose flux density is proportional to the current through the fixed coil. The moving coil is kept in between the fixed coil. When the current passes through the moving coil, a magnetic field is produced by this coil. The magnetic poles are produced in such a way that the torque produced on the moving coil deflects the pointer over the calibrated scale. This instrument works on AC and DC. When AC voltage is applied, alternating current flows through the fixed coil and moving coil. When the current in the fixed coil reverses, the current in the moving coil also reverses. Torque remains in the same direction. Since the current i1 and i2 reverse simultaneously. This is because the fixed and moving coils are either connected in series or parallel.

Errors in dynamometer type instruments

- **Frictional Error:** Since the coils are air-cored, therefore the magnetic field produced is of small strength. So they require a large number of ampere-turns to create necessary deflecting torque. This result in the heavy moving system. Therefore small torque-weight ratio. Thus the frictional losses in dynamo type instruments are somewhat larger as compared to other instruments.
- **Temperature errors:** Since the operation of dynamo type instrument required considerable power, self heating in these instrument is appreciable. The error due to self heating may be much as 1% of full scale deflection.
- **Error Due to Stray Magnetic field:** Since the operating magnetic field produced by the fixed coil. In these instruments is somewhat weaker in comparison to that in the instrument of other type. The operation of these instruments is more sensitive to the stray magnetic field.
- **Frequency error:** The change in frequency causes error
  - Due to change in reactance of operating coil.
  - Due to change in magnitude of Eddy current setup in the metal part of the instrument near to operating portion.

Advantages of Dynamometer type instrument

- As the instrument has Square Law response so can be used on both the dc as well as on AC.
- These instruments are free from hysteresis and Eddy current errors. It is because of absence of iron in the operating part of the instrument.
- Ammeter up to 10A and voltmeter up to 600V can be constructed with precision grade accuracy.
- Dynamo type voltmeter are useful for accurate measurement of rms value of voltage irrespective of waveform.
- Because of Precision grade accuracy and same calibration for DC and AC measurement instruments are used as transfer and calibration instruments.

Disadvantage of Dynamometer type instrument

- The scale is not uniform as the instrument uses Square Law response. These instruments have small torque-weight ratio so the friction error is considerable.
- Owing to heavy moving system friction losses in these instruments are somewhat more than those in other instruments.
- As a result of measures taken to reduce the frictional errors, their cost is more in comparison to moving iron and PMMC instruments. They are more sensitive to overload and mechanical impact and are to be handled with care.
- Adequate screening of the movements against the stray magnetic field is essential.
- The sensitivity of the instrument is typically very low due to poor deflecting torque. The sensitivity of dynamo type wattmeter is 10 to 30 per volt in comparison to the sensitivity of 20-kilo-ohm per volt in case of D'Arsnoval movement.
- The power consumption of this instrument is comparatively high because of their construction.

Ranges:

Ammeter: 1. With fixed and moving coil in series- 0/0.01A-0/0.05 A

2. With moving coil shunted or parallel connection- upto 0/30A.

Voltmeter: Upto 0-750 volts

EXTENSION OF RANGE OF INSTRUMENT BYTHE USE OF SHUNT AND MULTIPLIER

AMMETER CONNECTION

Shunts are used for the range extension of ammeters. A shunt is a low-value resistance having minimum temperature co-efficient and is connected in parallel with the ammeter whose range is to be extended. The combination is connected in series with the circuit whose current is to be measured.

The ratio of maximum current (with shunt) to the full-scale deflection current (without shunt) is known as the 'multiplying power' or 'multiplying factor' of the shunt.

Example: A moving coil ammeter reading up to 1 ampere has a resistance of 0.02 ohm. How could this instrument be adopted to read current up to 100 amperes**.**

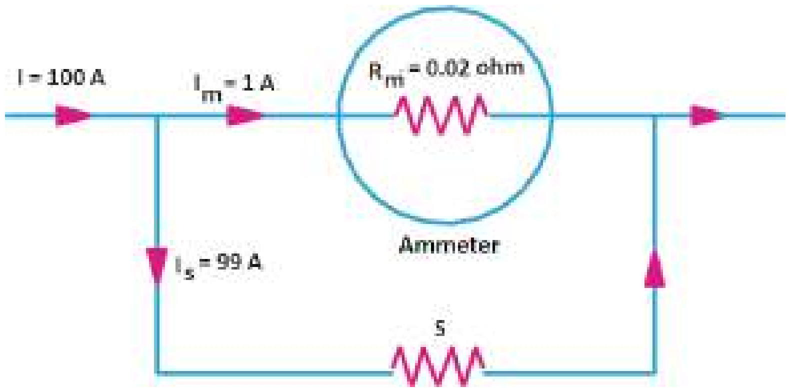
Solution: In this case,

Full-scale deflection current of the ammeter, $I_m = 1$ A

Line current to be measured, $I = 100$ A

Resistance of ammeter, $R_m = 0.02$ ohm

Let, the required shunt resistance = S



As seen from Figure, the voltage across the instrument coil and the shunt resistance is the same since
both are joined in parallel.

$\therefore I_m R_m = SI_s = S(I - I_m)$

or **S =** $I_m R_m/(I - I_m)$

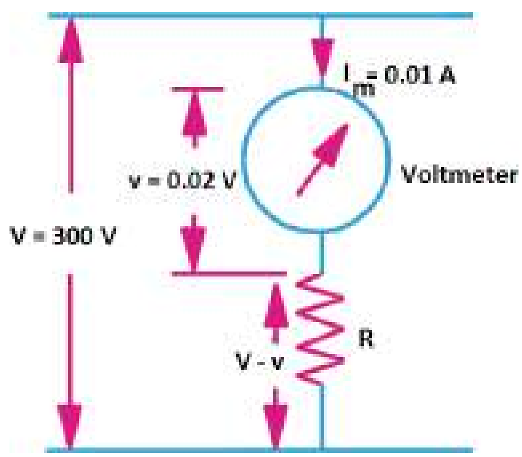
$= 1 \times 0.02/(100 - 1) = 0.02/99 = 0.000202$ **Ans**.

VOLTMETER CONNECTION

Multipliers are used for the range extension of voltmeters. The multiplier is a non-inductive high-value resistance connected in series with the instrument whose range is to be extended. The combination is connected across the circuit whose voltage is to be measured.

Example: A moving coil voltmeter reading upto 20 mV has a resistance of 2 ohms. How this instrument can be adopted to read voltage upto 300 volts.

Solution: In this case,
Voltmeter resistance, $R_m = 2$ ohm
Full-scale voltage of the voltmeter, $v = R_m I_m = 20$ mV $= 0.02$ V
Full-scale deflection current, $I_m = v/R_m = 0.02/2 = 0.01$ A
Voltage to be measured, $V = 300$ V
Let the series resistance required $= R$



Then as seen from figure, the voltage drop across R is V – ν

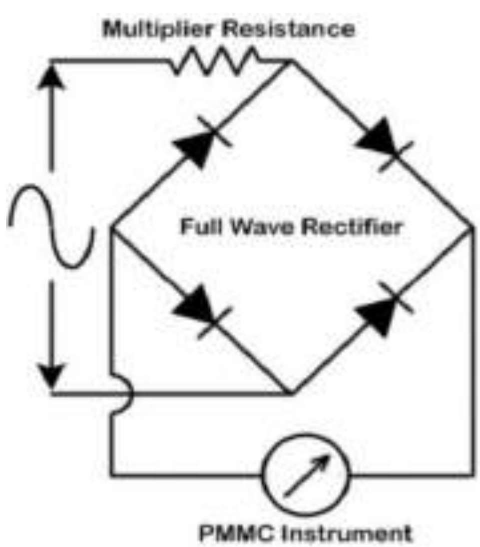
R I_m = V – ν

or **R = (V – v)/I_m**

or R = (300 – 0.02)/0.01 = 299.98/0.01 = 29998 ohms **Ans.**

RECTIFIER TYPE INSTRUMENT

A **rectifier type instrument** measures alternating electrical signal by means of D.C measuring instrument. As the name implies, this instrument first rectifies an A.C signal to D.C then measures. Although it measures the rectified A.C signal (D.C signal), but the scale of the instrument is calibrated for A.C. The sensitivity of D' Arsonval instrument is quite high. But a D'Arsonval instrument can only measure to D.C. So, to utilize the sensitivity of D'Arsonval movement for A.C, we use a **rectifier type instrument**.



Rectifier Elements for Rectifier Type Instrument

To convert A.C. to D.C, a rectifier instrument must have some rectifier elements.

At the low-frequency range, the instrument uses copper oxide or selenium cells for rectification purpose. Again at higher frequencies, we use germanium or silicon diodes.

The copper oxide rectifier element has a peak inverse voltage (PIV) about 2 volts. On the other hand, the selenium element has PIV of 10 volts. Also, both of these rectifier elements have very low current handling capacity. Therefore, for rectification purpose, these elements have become obsolete in the modern era.

Besides it, a germanium diode has a peak inverse voltage (PIV) of about 300 volts. In addition to that, the current carrying capacity of a germanium diode is about 100 mA. Also, PIV of silicon diode is about 1000 volts with a current rating of 5000 mA. Therefore these germanium and silicon diodes have become most suitable choice as rectifier elements for the purposes.

Advantages of Rectifier Type Instrument

1. This instrument can measure an electrical signal of very low frequency to radio frequency.
2. The instrument is also capable of measuring electrical signals up to several mega Hz.
3. The sensitivity of the instrument is much higher than that of any other type of A.C measuring instruments. Actually, for achieving very high sensitivity in A.C measurements, we use **rectifier type instruments**.

Characteristic of Rectifier Type Instrument

1. A rectifier type instrument is an economical and suitable mean of A.C measurement.
2. It can measure the electrical signal at an audio frequency range.
3. The sensitivity of the instrument is much higher than a commonly used A.C measuring instrument. It has been found that the sensitivity of a general rectifier type instrument is around 50 times higher than of a dynamometer type instrument or a moving iron type instrument.
4. Generally, D' Arsonval movement is uniform; therefore a rectifier type instrument generally has a linear scale.
5. The sensitivity of the instrument is in order of 1000 to 2000 ohms/volt.
6. The power consumption of a rectifier type instrument is high because of the resistance of the rectifier elements.
7. Shunting of the rectifier instrument is not practical because the resistance of rectifier elements changes with the temperature and the current flowing through it.
8. The rectifier type instrument is capable of measuring a very tiny current of microampere range. Again it can also be capable of measuring current in the milliampere range. But we do not construct a rectifier type instrument beyond for the current of 15mA. Because it requires special sized rectifier elements and the size of the instruments becomes impracticality large.

INDUCTION TYPE INSTRUMENT

The operation of Induction type instruments depends on the production of torque due to reaction between two magnetic fluxes having some phase difference or reaction between flux of an AC magnet and the eddy current induced by this flux. This instrument having an aluminum disc (or aluminum drum) in the magnetic field. Hence, the changing flux links with the aluminum disc. As a result, the flux induces an eddy current on the disc. This eddy current interacts with the flux which has induced it. Consequently, there is a mechanical torque acting on the disc. This mechanical torque rotates the disc. These type of instruments are used only for AC measurements.

Torque in Induction Type Instrument

So, the torque depends on two factors. The first one is the strength of the field of the electromagnet. The second one is the value of eddy current on the disc. Of course, the torque is proportional to the strength of the magnetic field. Also, it is proportional to the eddy current. Again, the strength of the magnetic field depends on the current of the electromagnet. On the other hand, the value of eddy current depends on the strength of the magnetic field. So, we can say, the value of eddy current also depends on the current of the electromagnet.

So, the torque acting on the disc is directly proportional to the square of the current of the electromagnet. In an induction type instrument, we directly feed the measuring current into the coil of the electromagnet. Therefore, the deflecting torque is directly proportional to the square of the measuring current.

Let us consider the flux, produced by the electromagnet is

$$\phi = \phi_m \sin\theta$$

The phase angle between that flux and induced eddy current is α. Hence, we can write the expression of the eddy current as

$$i = I_m \sin(\theta - \alpha)$$

Again, the instantaneous torque is directly proportional to the instantaneous eddy current and the flux. Hence, we can write,

$$T_{ins} \propto \phi i$$

So, the mean torque is as follows,

$$T_m \propto \frac{1}{\pi}\int_0^{\pi} \phi i d\theta$$

$$\Rightarrow T_m \propto \frac{1}{\pi}\int_0^{\pi} \phi_m \sin(\theta) I_m \sin(\theta-\alpha) d\theta$$

$$\Rightarrow T_m \propto \frac{\phi_m I_m}{2\pi}\int_0^{\pi} [cos\alpha - cos(2\theta-\alpha)] d\theta$$

$$\Rightarrow T_m \propto \frac{\phi_m I_m}{2\pi}\left|\theta \, cos\alpha - \frac{sin(2\theta-\alpha)}{2}\right|_0^{\pi}$$

$$\Rightarrow T_m \propto \frac{\phi_m I_m}{2\pi}(\pi \, cos\alpha) \Rightarrow T_m \propto \frac{\phi_m I_m}{2} cos\alpha$$

$$\Rightarrow T_m \propto \phi_{RMS} I_{RMS} \cos\alpha$$

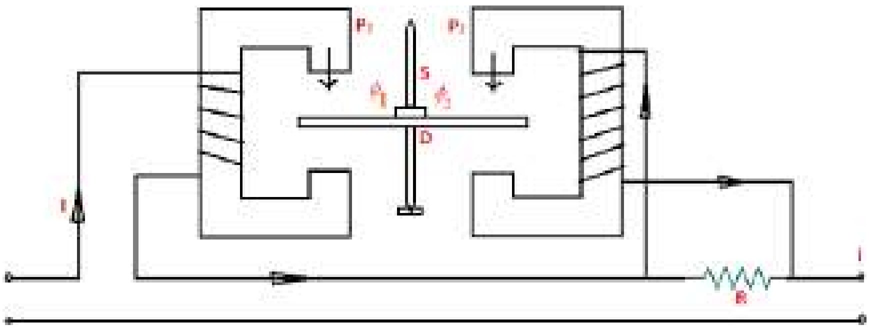
The above expression tells that the torque is zero if α is 90°. Hence to obtain resulting torque it is necessary to produce an eddy current which is either less than or more than 90° out of phase with flux ∅. a is the phase angle between the flux and eddy current.

So, there must be some means in induction type instrument to prevent this phase angle from being 90°. We can achieve this by two methods listed below.

1. Split-phase type and
2. Shaded pole type

1. SPLIT-PHASE TYPE INDUCTION TYPE INSTRUMENT:

In this arrangement, there are two AC magnets P1 and P2 connected in series. The winding in P2 is shunted by a resistance R. The current in the P2 winding lags with respect to the total current. This helps to develop the necessary phase angle a between the two fluxes. Eddy current damping is used in this type of instrument.

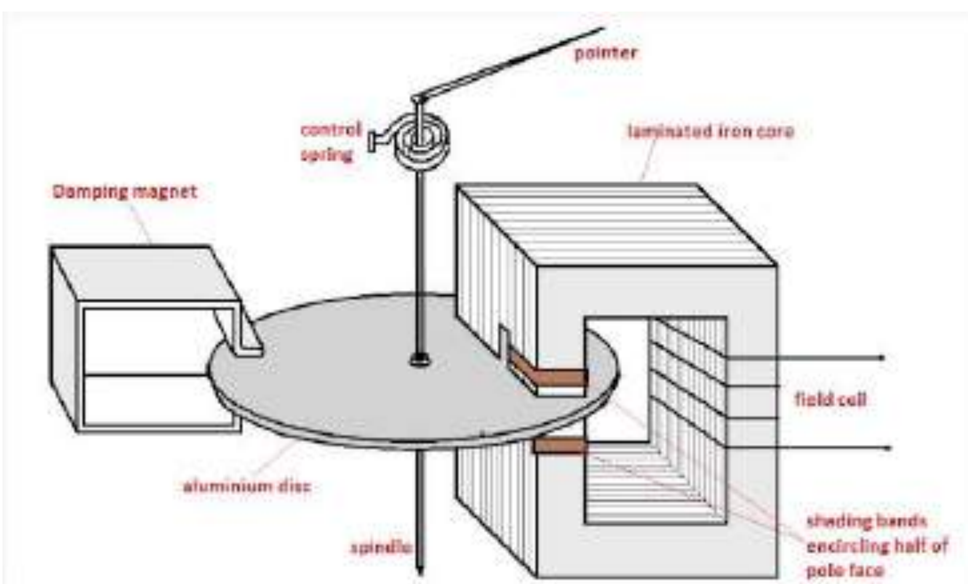


2. SHADDED POLE TYPE INDUCTION TYPE INSTRUMENT:

Shaded pole type induction instrument uses a single winding to produce flux. The flux produced by this winding is split up into two fluxes, having phase difference with respect to each other. The phase difference is usually 40 to 50 degrees and can be varied by varying the size of shading band. This is done by making a narrow slot in the poles of electromagnet. A copper strip is placed around the smaller of the two areas formed by t he slot. This copper shading band acts as a short circuited secondary winding.

The exciting coil is placed on the poles and a current proportional to current or voltage being measured is passed through it. An aluminium disc which is mounted on a spindle is inserted in the air gap of the electromagnet. The spindle carries a pointer and has a control sring attached to it. The controlling torque is provided by this spring only.

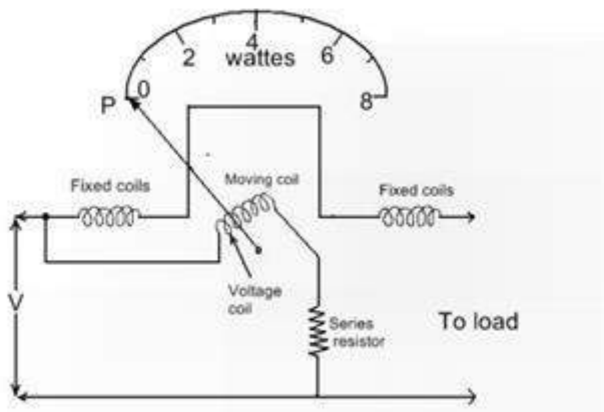
Damping is provided by a permanent magnet placed at the opposite side of the electromagnet, so that the disc can be used for production of both deflecting and damping torque.



CHAPTER-2

MEASUREMENT OF POWER

DYNAOMETER TYPE WATTMETER



Dynamometer type Wattmeter

A dynamometer type wattmeter primarily consists of two coils called fixed coil and moving coil. The fixed coil is splitted into two equal parts, which are placed parallel to each other. The two fixed coils are air-cored to avoid hysteresis effects when used on AC.

The fixed coil is connected in series with the load and carries the circuit current. It is, therefore, called the current coil. The moving coil is pivoted between the two parts of the fixed coil and is mounted on a spindle.

A pointer is attached to the spindle, which gives deflection. The moving coil is connected in parallel with the load and carries the current proportional to the voltage. It is, therefore, called the potential coil.Generally, a high resistance is connected in series with the moving coil to limit the current through it. By limiting the current, the moving coil is made lightweight, which in turn increases the sensitivity of the instrument.

The springs provide the controlling torque. They also serve the additional purpose of leading the current into and out of the moving coil. Air friction damping is employed in such instruments.

Dynamometer Type Wattmeter Working

We use the wattmeter for power measurements. Its current coil is connected in series with the load, carries the load current, and the potential coil, connected in parallel with the load, carries the current proportional to the voltage across the load.

The fixed coil produces a field F_m, and moving coil creates a field F_r. The field F_r tries to come in line with the main field F_m, which provides a deflecting torque on the moving coil.
Thus, the pointer attached to the spindle of the moving coil deflects. This deflection is controlled by the controlling torque produced by the springs. Also read Power Measurements in Three Phase Circuits.

Advantages and Disadvantages of Dynamometer Type Wattmeter

Advantages:

- It can be used both on AC and DC circuits.
- It has a uniform scale.
- We can obtain a high degree of accuracy through careful design.

Disadvantages:

- At low power factors, the inductance of the potential coil causes serious errors.
- The reading of the instrument may be affected by stray fields acting on the moving coil. To prevent it, magnetic shielding is provided by enclosing the instrument in an iron case.

Errors in Dynamometer Type Wattmeter

Errors in this type of wattmeter:

1. Error due to potential coil inductance: The inductance of the potential coil is liable to cause an error in the reading of the wattmeter. Because of this error, the wattmeter gives a high reading on lagging power factor and low reading on leading power factor.
The high non-inductive resistance connected in series with the coil swamps the phasing effect of the potential coil inductance.

2. Error due to power loss in the potential coil or current coil: Another possible error in the indicated power may be due to some voltage drop in the current coil or the current taken by the potential coil.
We can overcome this defect by using an additional compensating winding. This winding is connected in series with the potential coil and so placed that it produces a field in the opposite direction to that of the current coils.

3. Error due to eddy currents: The alternating field of fixed or current coil induces eddy currents in the solid metal parts which set up their own magnetic field. This alters the magnitude and phase of the magnetic field, causing deflection.

Thus an error is introduced in the instrument reading. To reduce this error, the solid metal parts are placed far away from the current coil as possible.

4. Error due to the stray magnetic field: The dynamometer type wattmeter has a relatively weak operating field; therefore, stray fields affect the reading of this instrument considerably and cause serious errors.
Hence, this type of instrument must be shielded against stray magnetic fields try using iron cases or providing thin iron shields over the working parts.

Range

- Current circuit 0 – 0.25A to 0 – I00A without employing CTs.
- Potential circuit 0 – 5 V to 0 – 750 V without using PTs.

INDUCTION TYPE WATTMETER

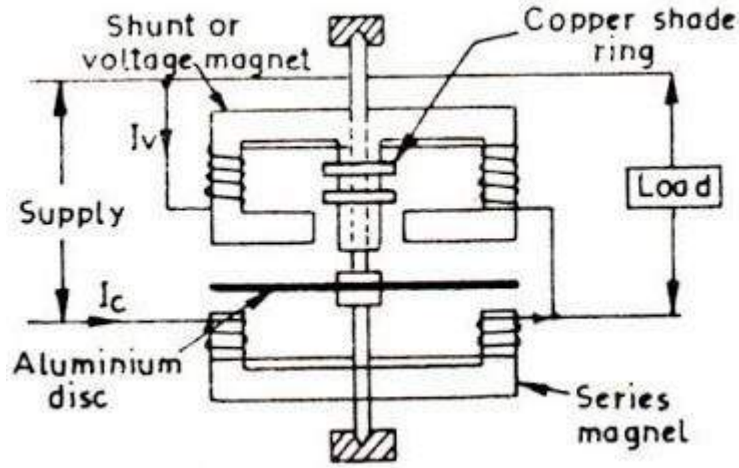
The induction type wattmeter is used to measure a.c power only.

Principle of Induction type wattmeter:

The principle of operation of an induction wattmeter is same as that of induction ammeters and voltmeters i.e. induction principle. However, it differs from induction ammeter or voltmeter in so far that separate two coils are used to produce the rotating flux in place of one coil with phase split arrangement.

Construction of Induction type wattmeter:

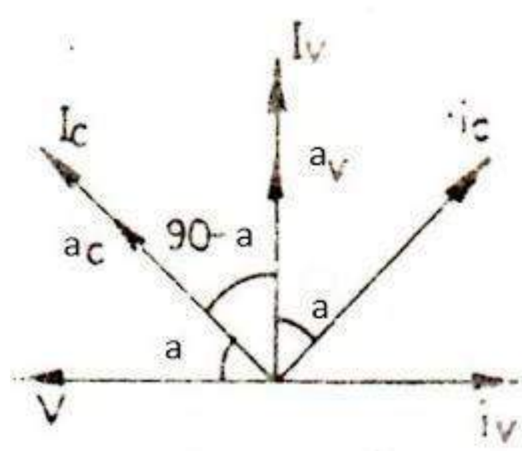
The principle parts of an induction wattmeter are as shown in the fig below. It consists of two laminated electromagnets. One electromagnet, called shunt magnet is connected across supply and carries current proportional to the applied voltage. The coil of this magnet is made highly inductive so that the current in it lags behind the supply voltage by 90 degrees. The other electromagnet, called series magnet is connected in series with supply and carries the load current. The coil of this magnet is made highly non inductive so that the angle of lag or lead is determined fully by the load.



A thin aluminium disc mounted on the spindle is placed in between the two magnets so that it cuts the fluxes of both the magnets. The controlling torque is provided by spiral springs. The damping is electromagnet and is usually provided by a permanent magnet embracing the aluminium disc. Two or more closed copper rings, called shading rings are provide on the central limb of the shunt magnet. By adjusting the position of these rings, the shunt magnet flux can be made to lag behind supply voltage by exactly 90degrees.

Working of Induction type wattmeter:

When the wattmeter is connected in the circuit to measure a.c power, the shunt magnet carries current proportional to the supply voltage and the series magnet carries the load current. The two fluxes produced by the magnets induce eddy currents in the aluminium disc. The interaction between the fluxes and eddy currents produce the deflecting torque on the disc, causing the pointer connected to the moving system to move over the scale.



Vector diagram

Deflecting torque of Induction type wattmeter:

let V = Applied voltage
Ic = Load current carried by the series magnet
Iv = Current carries by the shunt magnet
cos a = Lagging power factor of the load

The vector diagram of this wattmeter is shown in the fig below. The current Iv in the shunt magnet lags the applied voltage V by 90 degrees and so does the flux a_v produced by it. The current Ic in the series magnet is the load current and hence lags behind the applied voltage by a'. The flux ac produced by this current Ic is in phase with it. Therefore the two currents Ic in the current coil and Iv in the voltage coil and also corresponding fluxes a_v and ac are (90 - a') apart.
The flux ac induces the eddy currents iv in the aluminium disc which lags behind the flux by 90degrees. Similarly, flux ac induces eddy currents ic which again lags behind flux ac by 90 degrees.
Mean deflecting torque, Td proportional ac sin (90 - a)
Td proportional V I cos a
Td proportional a.c power

Since control is by springs, therefore
Tc proportional deflection
For steady deflected position, Td = Tc
Deflection proportional power
Hence, such instruments have uniform scale.

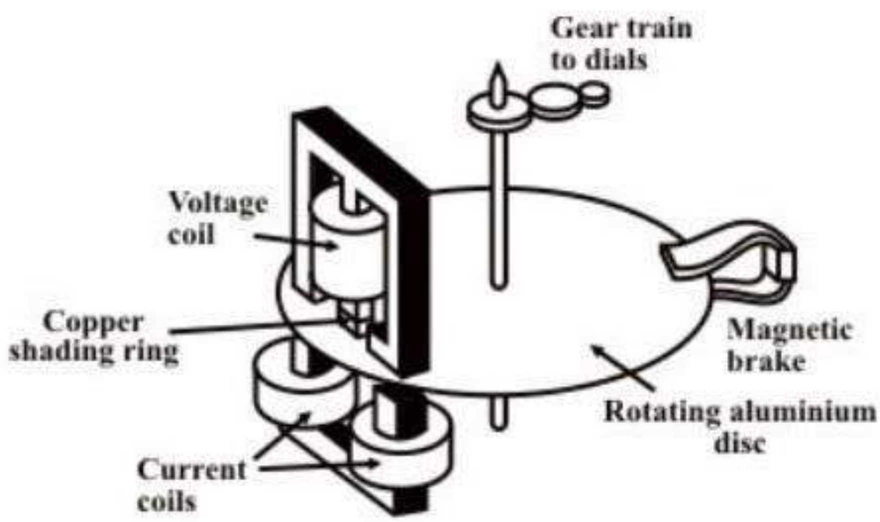
CHAPTER 4- MEASUREMENT OF ENERGY

SINGLE PHASE INDUCTION TYPE ENERGYMETERS:

Induction type energy meter consists of the following components:

(a) Driving system (b) Moving system (c) Braking system and (d) Registering system.

- Driving system: The construction of the electro magnet system is shown in Fig. 44.1(a) and it consists of two electromagnets, called "shunt" magnet and "series" magnet, of laminated construction.



Watt-hour meter.

Figure 1

A coil having large number of turns of fine wire is wound on the middle limb of the shunt magnet. This coil is known as "pressure or voltage" coil and is connected across the supply mains. This voltage coil has many turns and is arranged to be as highly inductive as possible. In other words, the voltage coil produces a high ratio of inductance to resistance. This causes the current, and therefore the flux, to lag the supply voltage by nearly 90^0. An adjustable copper shading rings are provided on the central limb of the shunt magnet to make the phase angle displacement between magnetic field set up by shunt magnet and supply voltage is approximately 90^0. The copper shading bands are also called the power factor compensator or compensating loop. The series electromagnet is energized by a coil, known as "current" coil which is connected in series with the load so that it carry the load current. The flux produced by this magnet is proportional to, and in phase with the load current.

- Moving system: The moving system essentially consists of a light rotating aluminium disk mounted on a vertical spindle or shaft. The shaft that supports the aluminium disk is connected by a gear arrangement to the clock mechanism on the front of the meter to provide information that consumed energy by the load. The time varying (sinusoidal) fluxes produced by shunt and series magnet induce eddy currents in the aluminium disc. The interaction between these two magnetic fields and eddy currents set up a driving torque in the disc. The number of rotations of the disk is therefore proportional to the energy consumed by the load in a certain time interval and is commonly measured in killowatt-hours (Kwh).

- Braking system: Damping of the disk is provided by a small permanent magnet, located diametrically opposite to the a.c magnets. The disk passes between the magnet gaps. The movement of rotating disc through the magnetic field crossing the air gap sets up eddy currents in the disc that reacts with the magnetic field and exerts a braking torque. By changing the position of the brake magnet or diverting some of the flux there form, the speed of the rotating disc can be controlled.

- Registering or Counting system: The registering or counting system essentially consists of gear train, driven either by worm or pinion gear on the disc shaft, which turns pointers that indicate on dials the number of times the disc has turned. The energy meter thus determines and adds together or integrates all the instantaneous power values so that total energy used over a period is thus known. Therefore, this type of meter is also called an "integrating" meter.

BASIC OPERATION:

Induction instruments operate in alternating-current circuits and they are useful only when the frequency and the supply voltage are approximately constant. The most commonly used technique is the shaded pole induction watt-hour meter, shown in fig.

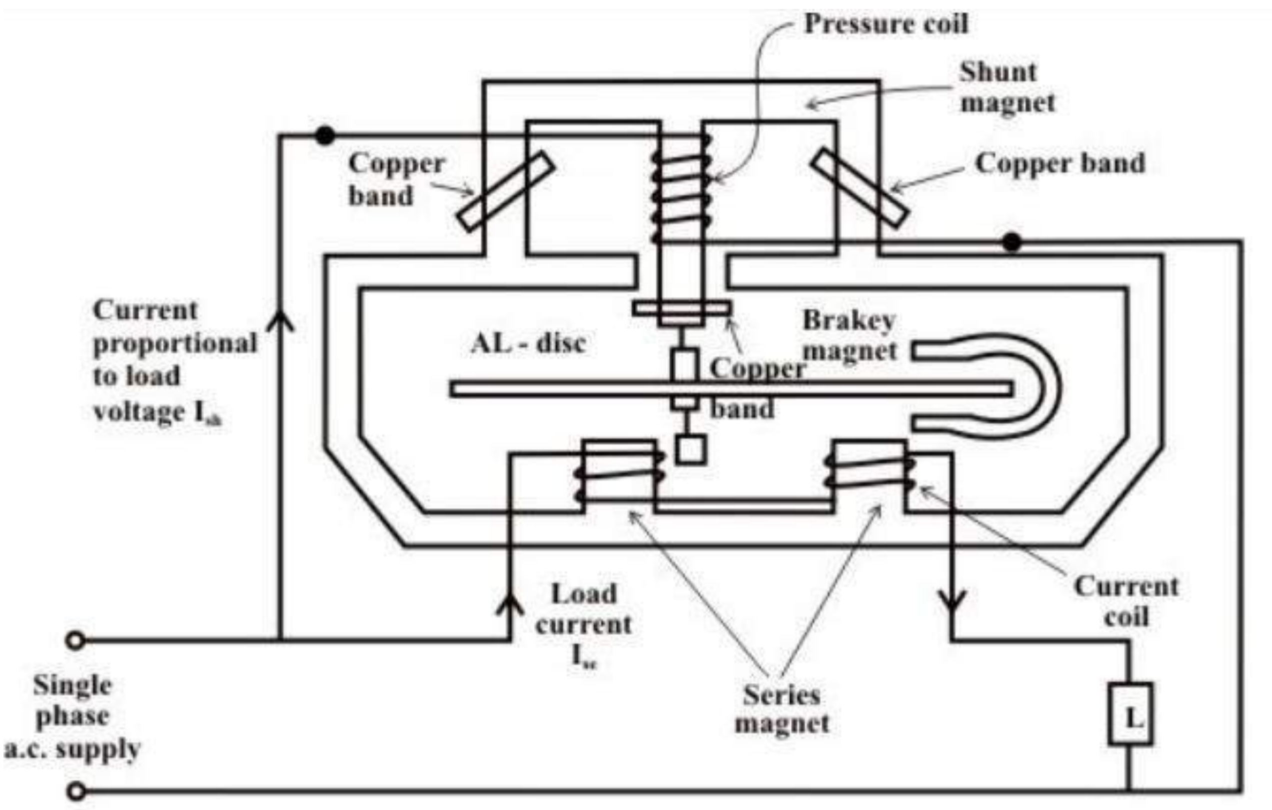


Figure 2

The rotating element is an aluminium disc, and the torque is produced by the interaction of eddy currents generated in the disc with the imposed magnetic fields that are produced by the voltage and current coils of the energy meter.

Let us consider a sinusoidal flux $\phi(t)$ is acting perpendicularly to the plane of the aluminium disc, the direction of eddy current i_e by Lenz's law is indicated in figure Fig. 2 . It is now quite important to investigate whether any torque will develope in aluminium disc by interaction of a sinusoidally varying flux $\phi(t)$ and the eddy currents i_e induced by itself.

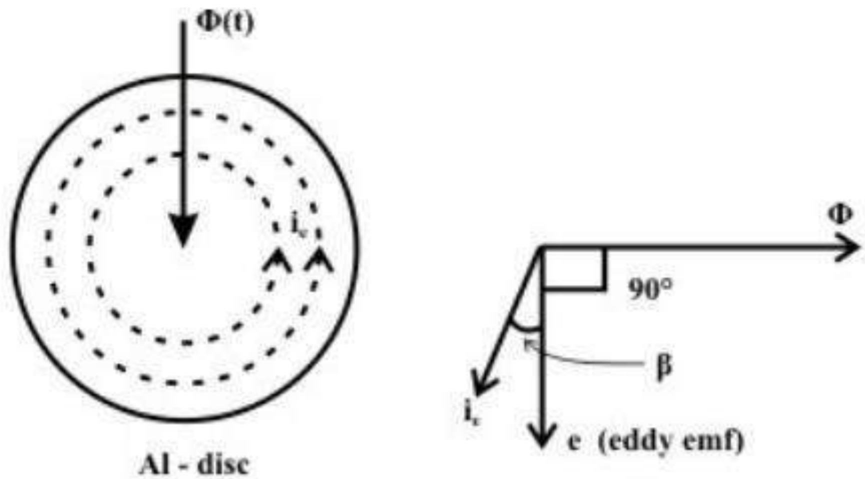


Fig. 3 : Eddy currents in aluminium disc due to time-varying flux.

$$\begin{aligned} T_{d(av.)} &\infty \ \phi\, I_e \cos(\angle \phi, I_e) = \phi\, I_e \cos(90^0 + \beta) \\ &\infty \ \phi\, I_e \sin(\beta) = 0 \end{aligned} \qquad (1)$$

where ϕ and I_e are expressed in r.m.s and $\beta = 0$ (because the reactance of the aluminium disc is nearly equal to zero). Therefore, the interaction of a sinusoidally varying flux $\phi(t)$ and its own eddy current i_e (induced) cannot produce torque any on the disc.

So in all induction instruments we have two fluxes produce by currents flowing in the windings of the instrument. These fluxes are alternating in nature and so they induce emfs in a aluminium disc or a drum provided for the purpose. These emfs in turn circulate eddy currents in the disc.

As in an energy meter instrument, we have two fluxes and two eddy currents and therefore two torques are produced by

i) first flux(ϕ_1) interacting with the eddy currents (I_{e2}) generated by the second flux (ϕ_2), and

ii) second flux (ϕ_2) interacting with the eddy currents (I_{e1}) induced by the first flux (ϕ_1).

In the induction type single phase energy meter, the flux produced by shunt magnet (pressure or voltage coil current) Φ_{sh} lags behind the applied voltage V by almost 90°. The flux ϕ_{se} is produced by the load current I and Φ_{se} is in the direction of I

Let the supply voltage $v(t)=V_{\max}\sin(\omega t)$ and load current $i(t)=I_{\max}\sin(\omega t-\theta)$. So, the fluxes are :

Let the supply voltage $v(t)=V_{\max}\sin(\omega t)$ and load current $i(t)=I_{\max}\sin(\omega t-\theta)$. So, the fluxes are :

(i) Flux generated by current coil

$$\Phi_{se}=k\,I_{\max}\sin(\omega t-\theta)=\Phi_{\max(se)}\sin(\omega t-\theta)$$

(ii) Flux generated by voltage coil

$$\Phi_{sh}=k'\int v(t)dt$$

$$=-k'\frac{V_{\max}}{\omega}\cos(\omega t)=\Phi_{\max(sh)}\sin(\omega t-90^0)$$

(note: $v(t)=\frac{1}{k'}\frac{d(\Phi_{sh})}{dt}$ and k and k' are constants.)

The eddy e.m.f , induced by flux Φ_{se} is

$$e_{se}\infty-\frac{\mathrm{d}}{\mathrm{dt}}(\Phi_{se})=-k\,I_{\max}\,\omega\cos(\omega t)$$

Eddy current generated in disc by the current coil

$$i_{se}\infty-\frac{k}{Z}I_{\max}\,\omega\cos(\omega t-\theta-\alpha)=\frac{k}{Z}I_{\max}\,\omega\sin(\omega t-(\theta+\alpha+90^0)),$$

where Z is the eddy current path impedance and α is the phase angle. In general, the angle $\alpha=\tan^{-1}\frac{X}{R}$ is negligible because $X=0$.

Also, note that

$$e_{sh} \propto -\frac{d}{dt}(\phi_{sh}) = -k'\frac{V_{max}}{\omega}\omega \sin(\omega t)$$

Eddy current generated in disc by the voltage coil

$$i_{sh} \propto -k'\frac{V_{max}}{Z}\sin(\omega t-\alpha) = k'\frac{V_{max}}{Z}\sin(\omega t+(180^0-\alpha))$$

The instantaneous torque on the disc is then proportional to

$$(\Phi_{sh}\, i_{se} - \Phi_{se}\, i_{sh}) = \frac{k\,k'}{Z}V_{max}\, I_{max}\left(\cos(\omega t)\,\cos(\omega t-\theta-\alpha) - \sin(\omega t-\theta)\,\sin(\omega t-\alpha)\right)$$

where Φ_{sh} is the flux generated by the voltage coil, Φ_{se} is flux generated by the current coil, i_{sh} is the eddy current produced in the disc by the voltage coil, and i_{se} is the eddy current produced in the disc by the current coil. The relative phases of these quantities are shown in fig.4

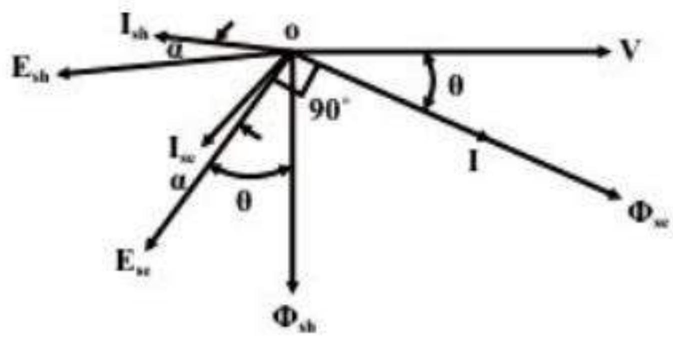


Fig. 4 : Phasor diagram of fluxes and eddy currents in watt-hour meter.

The flux generated by the current coil is in phase with the current and flux generated by the voltage coil is adjusted to be exactly in quadrature with the applied voltage by means of the copper shading ring on the voltage or shunt magnet. Theory of shaded pole is discussed in Appendix. The average torque acting upon the disc

$$T_{d(av)} \propto \frac{k\,k'}{Z} V_{\max} I_{\max} \frac{1}{2}\left(\cos(\theta+\alpha)+\cos(\theta-\alpha)\right)$$

$$\propto \frac{k\,k'}{Z} V_{\max} I_{\max} \cos\alpha \cos\theta = \left(\frac{2\,k\,k'}{Z}\cos\alpha\right) VI\cos\theta$$

$$\propto VI\cos\theta = \text{power in the circuit}$$

One can write average torque expression directly from the phasor diagram shown in fig.4

$$T_{d(av)} \propto \left[\Phi_{sh(rms)} I_{se} \cos(\angle\Phi_{sh(rms)}, I_{se}) - \Phi_{se(rms)} I_{sh} \cos(\angle\Phi_{se(rms)}, I_{sh})\right]$$

$$\propto \left[\Phi_{sh(rms)} I_{se} \cos(\theta+\alpha) - \Phi_{se(rms)} I_{sh} \cos(180+\alpha-\theta)\right]$$

$$\propto \left[k' V k \frac{I}{Z}\cos(\theta+\alpha) + k\, I\, k' \frac{V}{Z}\cos(\theta-\alpha)\right]$$

$$\propto \left(\frac{2\,k\,k'}{Z}\cos\alpha\right) VI\cos\theta$$

$$\propto VI\cos\theta = \text{power in the circuit}$$

where $\Phi_{sh}, \Phi_{se}, I_{sh}, I_{se}, V,$ *and* I are all expressed as r.m.s.

ERROS IN ENERGYMETER:

Assuming the supply voltage and frequency constant, the induction type energy may have the following errors:

i Speed error: Due to the incorrect position of the brake magnet, the braking torque is not correctly developed. This can be tested when meter runs at its full load current alternatively on loads of unity power factor and a low lagging power factor. The speed can be adjusted to the correct value by varying the position of the braking magnet towards the centre of the disc or away from the centre and the shielding loop. If the meter runs fast on inductive load and correctly on non-inductive load, the shielding loop must be moved towards the disc. On the other hand, if the meter runs slow on non-inductive load, the brake magnet must be moved towards the center of the disc.

ii Meter phase error: An error due to incorrect adjustment of the position of shading band results an incorrect phase displacement between the magnetic flux and the supply voltage (not in quadrature). This is tested with 0.5 p.f. load at the rated load condition. By adjusting the position of the copper shading band in the central limb of the shunt magnet this error can be eliminated.

iii Friction error: An additional amount of driving torque is required to compensate this error. The two shading bands on the limbs are adjusted to create this extra torque. This adjustment is done at low load (at about $1/4^{th}$ of full load at unity p.f.).

iv Creep: In some meters a slow but continuous rotation is seen when pressure coil is excited but with no load current flowing. This slow revolution records some energy. This is called the creep error. This slow motion may be due to (a) incorrect friction compensation, (b) to stray magnetic field (c) for over voltage across the voltage coil. This can be eliminated by drilling two holes or slots in the disc on opposite side of the spindle. When one of the holes comes under the poles of shunt magnet, the rotation being thus limited to a maximum of 180^0. In some cases, a small piece of iron tongue or vane is fitted to the edge of the disc. When the position of the vane is adjacent to the brake magnet, the attractive force between the iron tongue or vane and brake magnet is just sufficient to stop slow motion of the disc with full shunt excitation and under no load condition.

(v) Temperature effect: Energy meters are almost inherently free from errors due to temperature variations. Temperature affects both driving and braking torques equally (with the increase in temperature the resistance of the induced-current path in the disc is also increases) and so produces negligible error. A flux level in the brake magnet decreases with increase in temperature and introduces a small error in the meter readings. This error is frequently taken as negligible, but in modern energy meters compensation is adopted in the form of flux divider on the break magnet.

Energy meter constant K is defined as

$$K = \frac{\text{No. of revolutions}}{\text{kwh}}$$

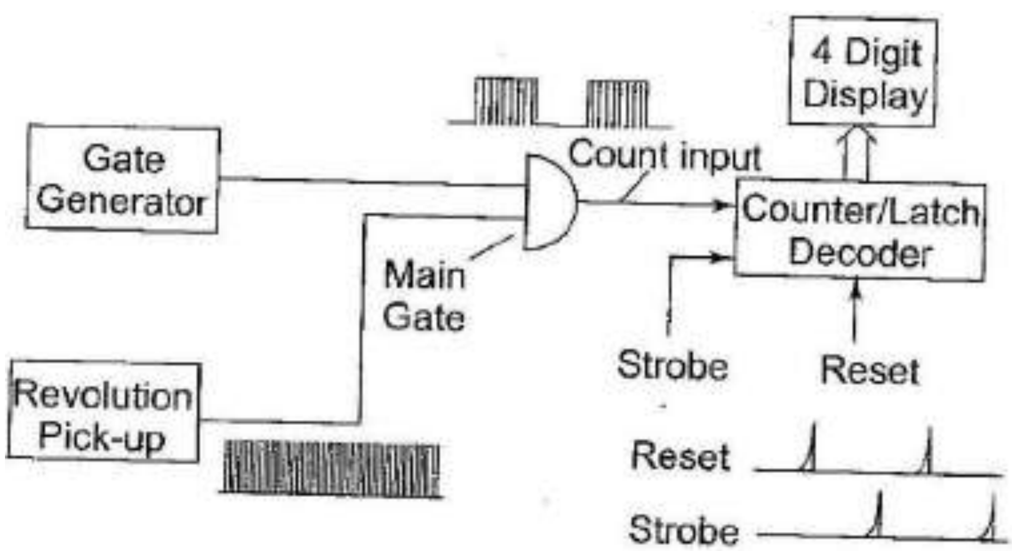
In commercial meters the speed of the disc is of the order of 1800 revolutions per hour at full load

CHAPTER-5

MEASUREMENT OF SPEED, FREQUENCY AND POWER FACTOR

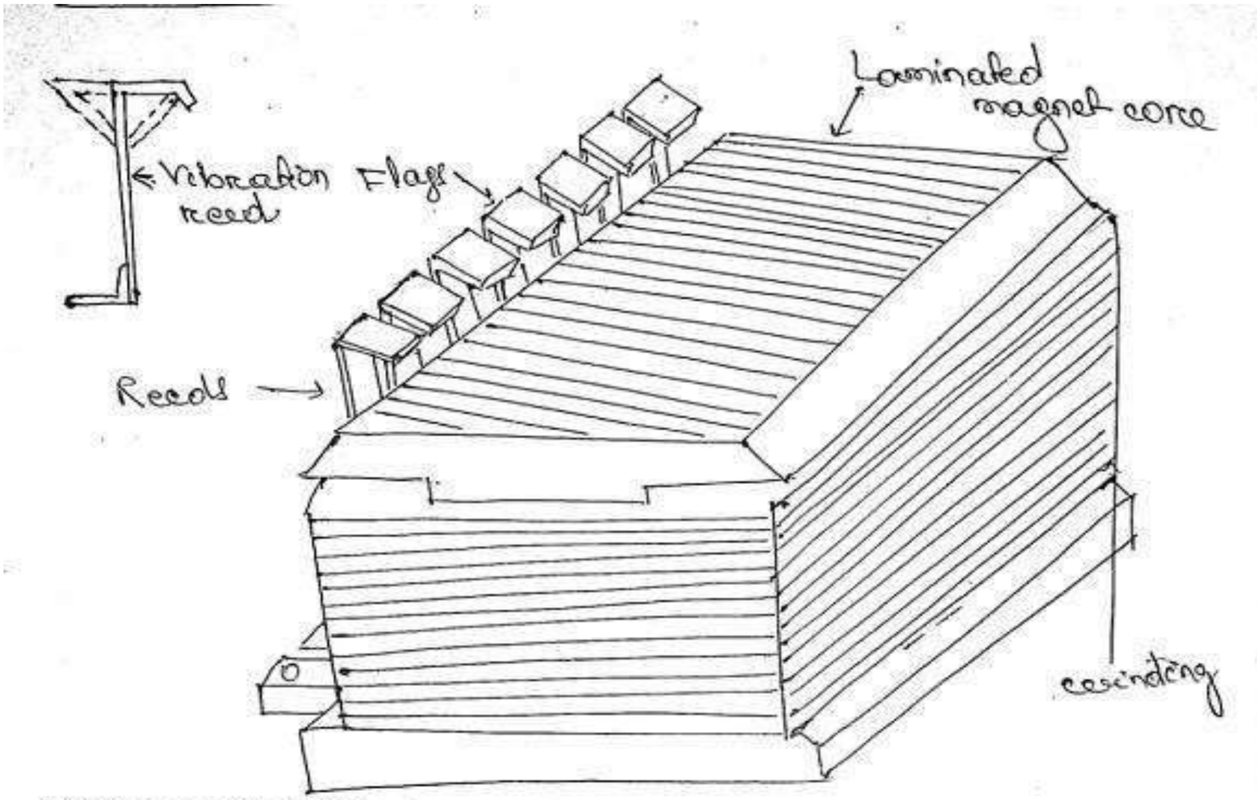
DIGITAL TACHOMETER

The technique employed in measuring the speed of a rotating shaft is similar to the technique used in a conventional frequency counter, except that the selection of the gate period is in accordance with the rpm calibration. Let us assume that the rpm of a rotating shaft is R. Let P be the number of pulses produced by the pickup for one revolution of the shaft. Therefore, in one minute the number of pulses from the pickup is R x P. Then, the-frequency of the signal from the pickup is (R x P)/60. Now, if the gate period is G s the pulses counted are (R x P x G)/60. In order to get the direct reading in rpm, the number of pulses to be counted by the counter is R. So we select the gate period as 60/ P, and the counter counts (Rx P x 60)/ 60P = R pulses and we can read the rpm of the rotating shaft directly. So, the relation between the gate period and the number of pulses produced by the pickup is G = 60/P. If we fix the gate period as one second (G= 1 s), then the revolution pickup must be capable of producing 60 pulses per revolution. Figure shows a schematic diagram of a digital tachometer.



Basic Block Diagram of a Digital Tachometer

MECHANICAL RESONANCE TYPE FREQUENCYMETER:



CONSTRUCTION:-

- The meter consists of number of thin steel strips, known as reeds. These reeds are arranged alongside and close to a electromagnet.
- The electromagnet is a laminated iron core. and Its coil is connected in series with a resistance across the supply circuit whose freq. is to be measured.
- Bottom portion of reeds are fixed and upper portion is kept free to vibrate. At the free end reeds are bent to form a flag. The reeds are painted white to distinguish them against black background.

- The reeds have either different dimensions or carry different weights or flags at their tops.
- The reeds are so designed and arranged that the natural freq. of one reed is differ from another by one or half cycle. So if the freq. meter has a range of 47 – 53 Hz then natural freq. of first reed will be 47 Hz and, 2nd reed be 47.5 Hz 3rd be 48 Hz and so on.

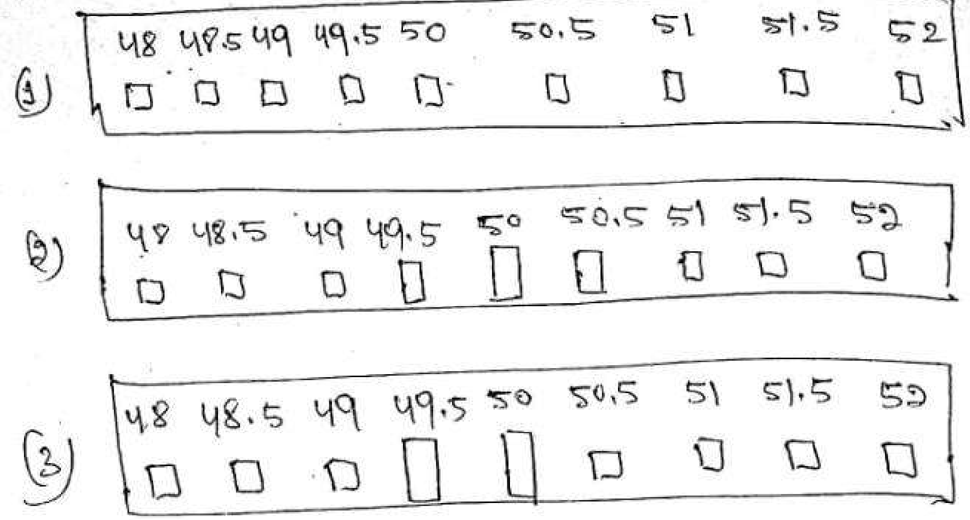
Operation :-

When the freq. meter is connected across the supply whose freq. is to be measured, the coil of electromagnet carries a current i. Due to the alternating field, a force of attraction is experienced between the reeds and the electromagnet, which is proportional to i² and ∴ this force varies at twice the supply frequency. Thus the force is exerted on the reeds varies every half cycle. As a result all the reeds will tend to vibrate but the reeds whose natural freq is equal to twice the supply freq. will vibrate with maxm amplitude. The vibration of other reeds are so slight as to be unobserved. The freq. is determined, by noting the scale reading opposite the reed that vibrate with maxm flags amplitude. If two adjacent reeds vibrate with equal amplitude then the supply freq. will be half way between the frequencies of two adjacent reeds.

The usual range of freq. meters of this type is about 6Hz (say from 47 Hz to 53 Hz)

(1) Show the condn when the meter is not connected to the supply.
(2) Show the condn when 50 Hz reed is vibrating with its maxm ~~ap~~ amplitude.
(3) Show the condn when the freq. is exactly midway between 49.5 Hz and 50 Hz.

(indication from vibrating reeds)



The range of the instrument may be doubled by polarizing its reeds. The polarizing may be done by using a dc winding in addition to the ac winding or by using a permanent magnet.

In the presence of alternating flux, the reeds are attracted two times in a cycle and the reed whose freq. is twice the supply freq. will responds, If the electromagnet is polarized by dc wdg in addition to ac wdg, the fields (ac and dc) will cancel each other in one half cycle, so the and reinforce each other in other half cycle, so the reeds will be attracted only once in a cycle. Thus a reed whose natural freq. is 100 Hz will respond to 50 Hz when the electromagnet is unpolarized and to 100 Hz when the electromagnet is polarized.

Advantage

The freq. to be measured is independent of waveform of supply voltage.

Disadvantage:

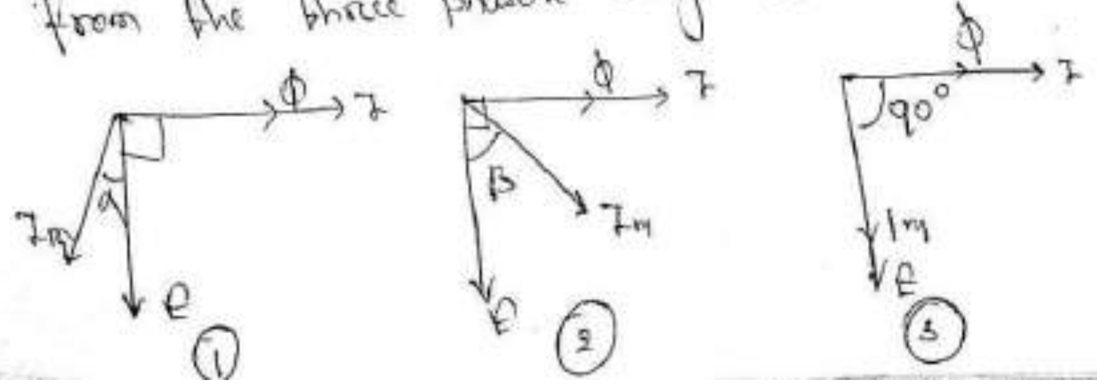
1. Amplitude of vibration depends upon the voltage and if voltage is too low to give appreciable amplitude of vibration, readings will be unreliable.
2. These are not for precision measurement of frequency since adjacent reeds have only difference of 0.5 Hz of natural frequency.

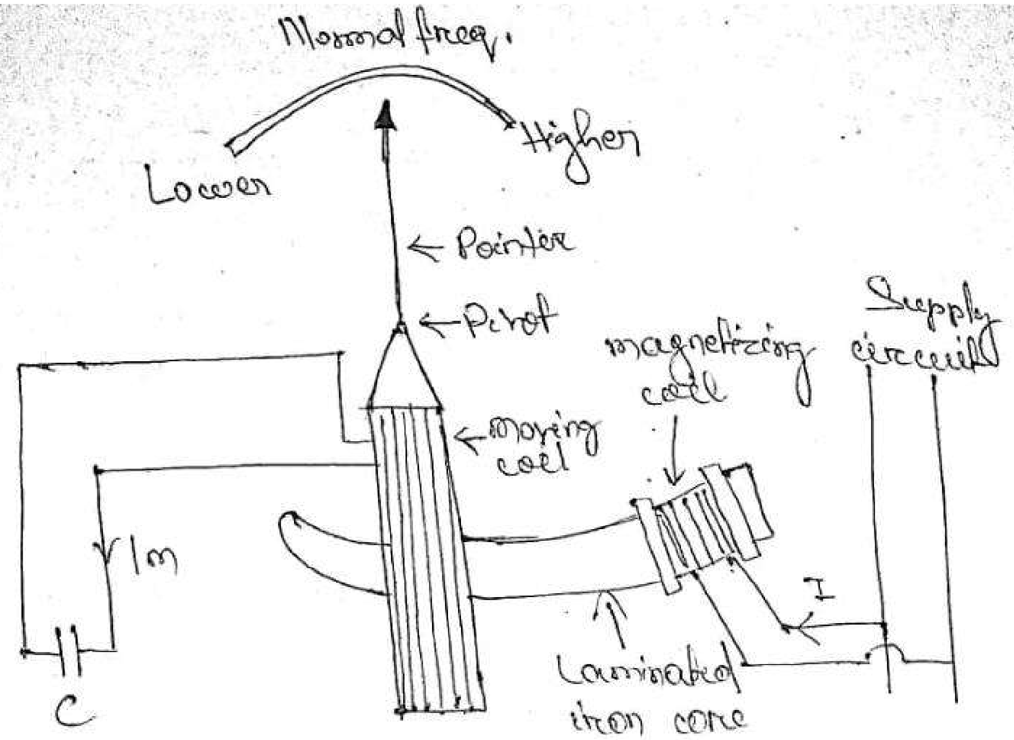
Electrical Resonance Type Frequency meter

(1) Ferrodynamic type Freq. meter :-

Construction: It consists of fixed coil which is connected across the supply whose freq. is to be measured. This coil is called magnetizing coil. The magnetizing coil is mounted at one end of a laminated iron core of varying cross section (max^m near the end where magnetizing coil is mounted and minimum at other end). A moving coil and with a pointer is pivoted over the iron core so that it can move freely over the iron core. The moving coil is connected across a capacitor C.

Working: The op^n of the instrument can be understood from the three phasor diagram.





When the magnetising coil is connected across the supply circuit, current I flows through it and ~~a flux φ~~ produces a flux φ, (in phase with its current I), in the iron core. This flux induces an emf E in the moving coil which lags behind the flux φ by 90°. This emf E circulate a current Im in the moving coil. The phase of this current Im depends on the inductance of the moving coil and the capacitance C.

phasor

1. when $X_L > X_C$, Im lags behind E by α.

   $\therefore T_d \propto I_m \cos(90+\alpha)$

2. when $X_L < X_C$, Im leads E by β

   $\therefore T_d \propto I_m \cos(90-\alpha)$

3. when $X_L = X_C$, Im is in phase with E (resonance cond^n)

   $\therefore T_d \propto I_m \cos 90 = 0$

For a fixed freq. the capacitive reactance is constant but the X_L of moving coil is not constant. This is bec. its X_L depends on the posn occupied by MC on the iron core. X_L becomes maxm when MC is closer to magnetizing coil and minimum when it is at the other end.

Hence the moving coil is pulled towards the magnetizing coil until $X_L = X_C$ or $\omega L = \frac{1}{\omega C}$ and the torque is zero i.e the circuit of moving coil is in resonance.

The value of C is so chosen that the moving coil take up a mean posn at its normal value. If frequency increases $\frac{1}{\omega C}$ or X_C decreases and X_L i.e ωL increases. Thus the circuit becomes largely inductive which produce deflecting torque which tries to pull the moving coil to an equilibrium posn i.e a posn where $X_L = X_C$. This can be obtained by moving the moving coil away from magnetizing coil. The coil moves farther 'on' to the core if freq. decreases.

Advantages Great sensitivity can be achieved ~~with the use~~ if the inductance of moving coil changes sharply with the variation of its posn on the core.

CHAPTER -6

MEASUREMENT OF RESISTANCE, INDUCTANCE, CAPACITANCE

Explain the working of Wheatstone Bridge(Measurement of Resistance)

For measuring accurately any electrical resistance Wheatstone bridge is widely used. There are two known resistors, one variable resistor and one unknown resistor connected in bridge form as shown below. By adjusting the variable resistor the electric current through the Galvanometer is made zero. When the electric current through the galvanometer becomes zero, the ratio of two known resistors is exactly equal to the ratio of adjusted value of variable resistance and the value of unknown resistance. In this way the value of unknown electrical resistance can easily be measured by using a Wheatstone Bridge.

Wheatstone Bridge Theory

The general arrangement of **Wheatstone bridge circuit** is shown in the figure below. It is a fourarms bridge circuit where arm AB, BC, CD and AD are consisting of electrical resistances P, Q, Sand R respectively. Among these resistances P and Q are known fixed electrical resistances andthese two arms are referred as ratio arms. An accurate and sensitive Galvanometer is connectedbetween the terminals B and D through a switch S2. The voltage source of this Wheatstonebridge is connected to the terminals A and C via a switch S1 as shown. A variableresistor S isconnected between point C and D. The potential at point D can be varied by adjusting the valueof variable resistor. Suppose electric current I1 and electric current I2 are flowing through thepaths ABC and ADC respectively. If we vary the electrical resistance value of arm CD the valueof electric current I2 will also be varied as the voltage across A and C is fixed. If we continue toadjust the variable resistance one situation may comes when voltage drop across the resistor Sthat is I2.S is becomes exactly equal to voltage drop across resistor Q that is I1.Q. Thus thepotential at point B becomes equal to the potential at point D hence potential difference betweenthese two points is zero hence electric current through galvanometer is nil. Then the deflection inthe galvanometer is nil when the switch S2 is closed.

Now, from **Wheatstone bridge circuit**

$$\boxed{\text{current } I_1 = \frac{V}{P + Q}}$$

and

$$\boxed{\text{current } I_2 = \frac{V}{R + S}}$$

Now potential of point B in respect of point C is nothing but the voltage drop across the resistorQ and this is

$$\boxed{I_1.Q = \frac{V.Q}{P + Q} \text{-------------------(i)}}$$

Again potential of point D in respect of point C is nothing but the voltage drop across theresistor S and this is

$$\boxed{I_2.S = \frac{V.S}{R + S} \text{------------------(ii)}}$$

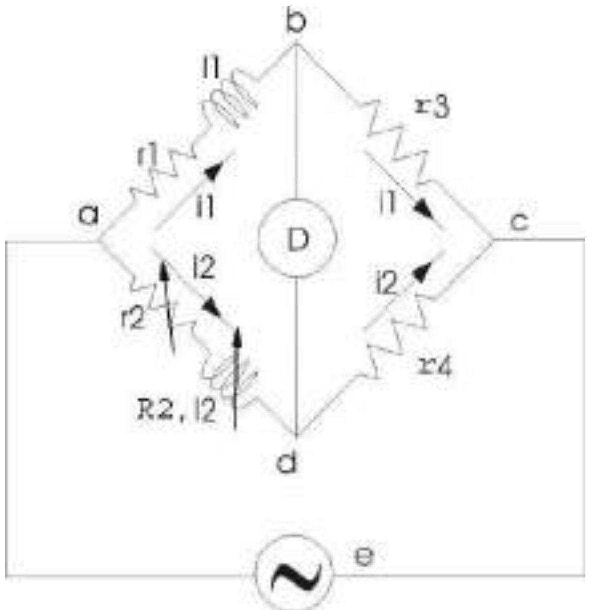
Equating, equations (i) and (ii) we get,

$$\frac{V.Q}{P + Q} = \frac{V.S}{R + S} \Rightarrow \frac{Q}{P + Q} = \frac{S}{R + S}$$
$$\rightarrow \frac{P + Q}{Q} = \frac{R + S}{S} \rightarrow \frac{P}{Q} + 1 = \frac{R}{S} + 1 \rightarrow \frac{P}{Q} = \frac{R}{S}$$
$$\rightarrow R = SX\frac{P}{Q}$$

Here in the above equation, the value of S and P ⁄ Q are known, so value of R can easily be determined. The electrical resistances P and Q of the Wheatstone bridge are made of definite ratio such as 1:1; 10:1 or 100:1 known as ratio arms and S the rheostat arm is made continuously variable
from 1 to 1,000 Ω or from 1 to 10,000 Ω.

MAXWELLS BRIDGE:

This bridge is used to find out the self inductor and the quality factor of the circuit. As it is basedon the bridge method (i.e. works on the principle of null deflection method), it gives veryaccurateresults. Maxwell bridgeis an AC bridge so before going in further detail let us know moreaboutthe ac bridge.Let us now discuss Maxwell's inductor bridge. The figure shows the circuit diagram ofMaxwell's inductor bridge.

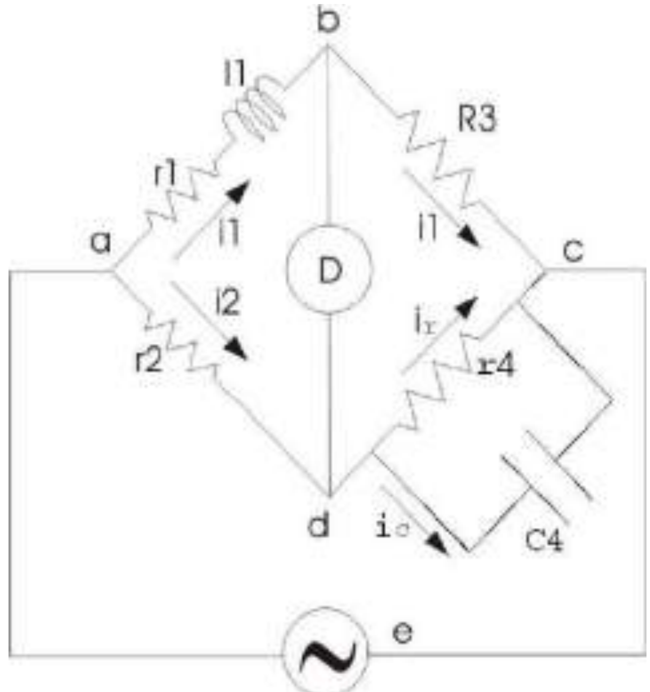


Maxwells Bridge

In this bridge the arms bc and cd are purely resistive while the phase balance depends on the arms ab and ad.
Here l1 =Unknown inductor of r1.
l2 =Variable inductor ofresistanceR2.
r2 =variable electricalresistance
As we have discussed in ac bridge according to balance condition, we have at balance point
We can vary R3 and R4 from 10 ohms to 10,000 ohms with the help of resistance box.

MAXWELL'S INDUCTANCE CAPACITANCE BRIDGE

In this Maxwell Bridge, the unknown inductor is measured by the standard variable capacitor. Circuit of this bridge is given below.



Maxwell's Inductance Capacitance Bridge

Advantages of Maxwell's Bridge

(1) The frequency does not appear in the final expression of both equations, hence it isindependent of frequency.

(2) Maxwell's inductor capacitance bridge is very useful for the wide range of measurementof inductor at audio frequencies.

Disadvantages of Maxwell's Bridge

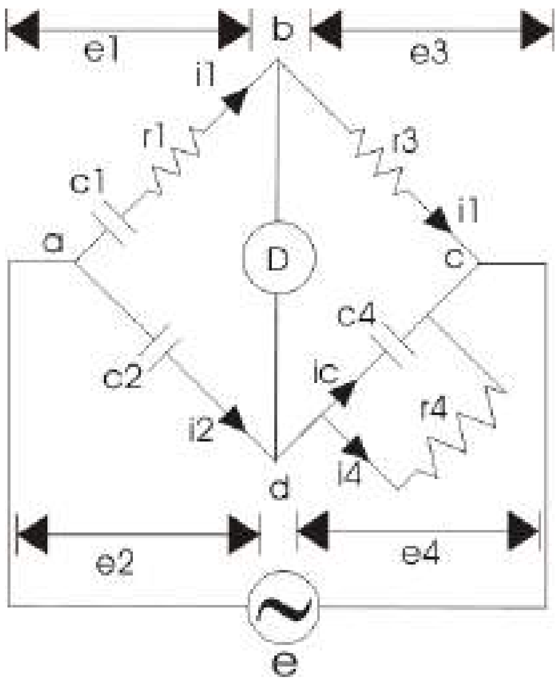
(1) The variable standard capacitor is very expensive.

(2) The bridge is limited to measurement of low quality coils ($1 < Q < 10$) and it is also unsuitable

for low value of Q (i.e. $Q < 1$) from this we conclude that a Maxwell bridge is used suitable only for medium Q coils.

SCHERING BRIDGE THEORY

This bridge is used to measure to the capacitance of the capacitor, dissipation factor andmeasurement of relative permittivity. Let us consider the circuit of Schering bridge as shownbelow



Schering Bridge

Here, c1 is the unknown capacitance whose value is to be determined with series electricalresistance r1.

c2 is a standard capacitor.

c4 is a variable capacitor.

r3 is a pure resistor (i.e. non inductive in nature).

And r4 is a variable non inductive resistor connected in parallel with variable capacitor c4.

Now the supply is given to the bridge between the points a and c. The detector is connectedbetween b and d. From the theory of ac bridges we have at balance condition

$$z_1 z_4 = z_2 z_3$$

Substituting the values of z_1, z_2, z_3 and z_4 in the above equation, we get

$$\left(r_1 + \frac{1}{j\omega c_1}\right)\left(\frac{r_4}{1+j\omega c_4 r_4}\right) = \frac{r_3}{j\omega c_2}$$

$$\left(r_1 + \frac{1}{j\omega c_1}\right) r_4 = \frac{r_3}{j\omega c_2}(1 + j\omega c_4 r_4)$$

$$r_1 r_4 - \frac{j r_4}{\omega c_1} = -\frac{j r_3}{\omega c_2} + \frac{r_3 r_4 c_4}{c_2}$$

Equating the real and imaginary parts and the separating we get,

$$r_1 = \frac{r_3 c_4}{c_2}$$

$$c_1 = c_2 \frac{r_4}{r_3}$$

Application:

This bridge is used to measure to the capacitance of the capacitor, dissipation factor andmeasurement of relative permittivity.

TRANSDUCERS AND SENSORS

METHOD OF SELECTING TRANSDUCERS

While selecting the proper transducer for any applications, or ordering the transducers the following specifications should be thoroughly considered.

1) Ranges available
2)Squaring System
3)Sensitivity
4) Maximum working temperature
5) Method of cooling employed
6) Mounting details
7) Maximum depth
8) Linearity
and hysteresis
9) Output for zero input
10) Temperature co-efficient of zero drift
11) Natural Frequency.

ADVANTAGES OF ELECTRICAL TRANSDUCERS

1. Very small power is required for controlling the electrical or electronic system
2. The electrical output can be amplified to any desired level
3. Mass inertia effects are reduced to minimum possible.
4. The size and shape of the transducers can be suitably designed to achieve the optimum weight and volume

5. The output can be indicated and recorded remotely at a distance from the sensing medium .
6. The outputs can be modified to meet the requirements of the indicating or controlling equipment.

RESISITIVE TRANSDUCERS

The resistance of a conductor is expressed by a simple equation that involves a few physical quantities . The relationship is given by

$R= \rho L/A$

Where , R= resistance, Ω

ρ = Resistivity of conductor materials, Ω-m

L= Length of conductor, m

A = Cross sectional area of the conductor,m2

Any method of varying one of the quantities involved in the above relationship can be the designedbasis of an electrical resistance transducer. There are a number of ways in whichresistance can bechanged by a physical phenomenon.The translational and rotationalpotentiometer which work on the basis of change in the value ofresistance with change in length of the conductor can be used for measurement of translational orrotatory displacements.

The resistivity of materials changes with the change of temperature thus causing a change ofresistance. This property may be used for measurement of temperature.In a resistance transducer an indication of measured physical quantity is given by a change in theresistance. It may be classified as follows

1. Mechanically varied resistance - POTENTIOMETER
2. Thermal resistance change – RESISTANCE THERMOMETER
3. Resistivity change - RESISTANCE STRAIN GAUGE

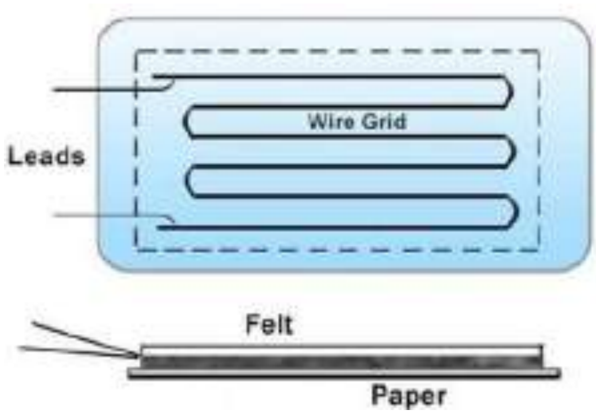
STRAIN GAUGE

INTRODUCTION

When a metal conductor is stretched or compressed , its resistance changes on account of the fact thatboth length and diameter of conductor change . The value of resistivity of conductor also changes.When it is strained it's property is called piezo-resistance .Therefore , resistance strain gauges arealso known as piezo- resistive gauges .The strain gauge is a measurement transducer for measuring strain and associated stress inexperimental stress analysis.

TYPES

Four types of Strain gauges are :
1. Wire –wound strain gauge
2. Foil-type strain gauge
3. Semiconductor strain gauge
4. Capacitive strain gauge.



WORKING PRINCIPLE

Strain gauges work on the principle that the resistance of a conductor or a semiconductor changeswhen strained .This property can be used for measurement of displacement, force and pressure .When a strain gauge is subjected to tension (positive strain) it's length increases while it's crosssectional area decreases. Since the resistance of a conductor is proportional to it's length and inverselyproportional to it's area of cross section, The resistance of the gauge increases with positive strain .Strain gauges are most commonly used in wheat –stone bridge circuits to measure the change ofresistance of grid of wire for calibration proposes; the 'GAUGE FACTOR' is defined as the ratio ofper unit change in resistance to per unit change in length.

i.e , Gauge factor (Gf) = $\Delta R/R \div \Delta L/L$

Where, ΔR = corresponding change in resistance, R

ΔL = Change in length per unit length, L

$R= \rho L/A$

Where, R= resistance, Ω

ρ = Resistivity of conductor materials, Ω-m

L= Length of conductor, m

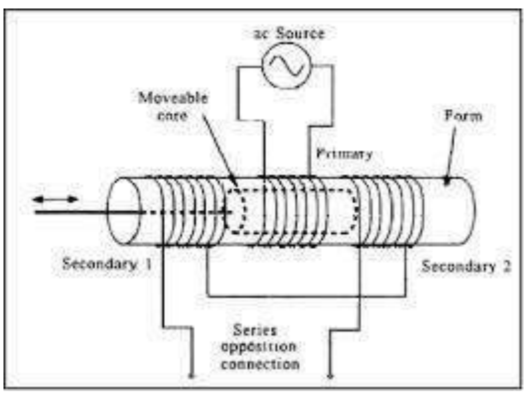
A = Cross sectional area of the conductor, m^2

L.V.D.T

LVDT is a passive inductive transducer and is commonly employed to measure force(or weight,pressure and acceleration etc. Which depend on force)in terms of the amount and direction ofdisplacement of an object.

WORKING PRINCIPLE

When the core is in the centre (called reference position) the induced voltages E1 and E2 are equal andopposite. Hence they cancel out and the output voltages V0 is zero.When the external applied force moves the core towards the coil S2 ,E2 is increased but E1 isdecreased in magnitude though they are still antiphase with each other. The net voltage available is(E2-E1) and is in phase with E2.



Similarly , When movable core moves towards coil S1, E1>E2 and Vo = E1-E2 and is in phase with E1.

ADVANTAGES

1. It gives a high output and therefore many a times there is no need for intermediate amplification devices.
2. The transducer possess a high sensitivity as high as 40V/mm

3. It shows a low hysteresis and hence repeatability is excellent under all conditions.
4. Most of the LVDTs consume a power of less than 1W.
5. Less friction and less noise

DISADVANTAGES

1. These transducers are sensitive to stray magnetic fields but shielding is possible .This is done byproviding magnetic shields with longitudinal slots.
2. Relatively large displacements are required for appreciable differential output.
3.Several times, the transducer performance is affected by vibrations.

APPLICATIONS

1. Measurement of material thickness in hot strip or slab steel mills
2. In accelerometers.
3. Jet engine controls in close proximity to exhaust gases.

CAPACITIVE TRANSDUCER (PRESSURE)

A linear change in capacitance with changes in the physical position of the moving element may be used to provide an electrical indication of the element's position.

The capacitance is given by $C = KA/d$

where K = the dielectric constant
A = the total area of the capacitor surfaces
d = distance between two capacitive surfaces
C = the resultant capacitance.

From this equation, it is seen that capacitance increases (i) if the effective area of the plate is increased, and (ii) if the material has a high dielectric constant.

The capacitance is reduced if the spacing between the plates is increased.

Transducers which make use of these three methods of varying capacitance have been developed.

With proper calibration, each type yields a high degree of accuracy. Stray magnetic and capacitive effects may cause errors in the measurement produced, which can be avoided by proper shielding. Some capacitive dielectrics are temperature sensitive, so temperature variations should be minimised for accurate measurements.

A variable plate area transducer is made up of a fixed plate called Stator and a movable plate called the Rotor.

The rotor is mechanically coupled to the member under test. As the member moves, the rotor changes its position relative to the stator, thereby changing the effective area between the plates. A transducer of this type is shown in Fig. 13.28.

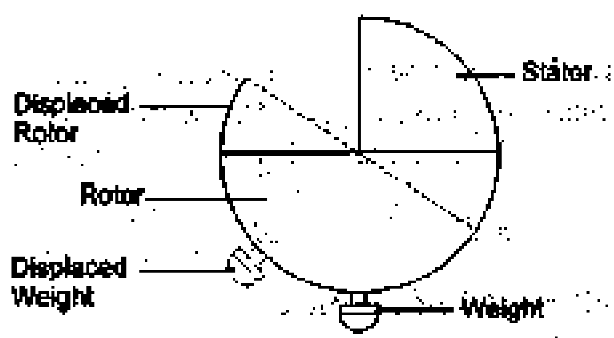


Fig. 13.28 Capacitive Transducer

Such a device is used to detect the amount of roll in an aircraft. As the aircraft rolls to the left, the plates moves to the relative position shown by dashed lines in Fig. 13.28 and the capacitance decreases by an amount proportional to the degree of roll. Similarly to the right. In this case the stator, securely attached to the aircraft, is the moving element. The weight on the rotor keeps its position fixed with reference to the surface of the earth, but the relative position of the plates changes and this is the factor that determines the capacitance of the unit.

Figure 13.29 shows a transducer that makes use of the variation in capacitance resulting from a change in spacing between the plates. This particular transducer is designed to measure pressure (in vacuum)

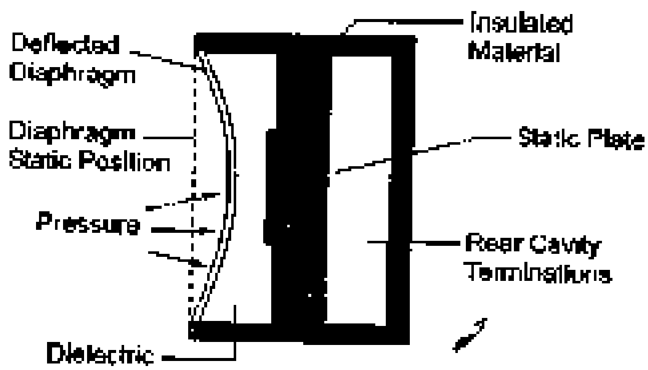


Fig. 13.29 Capacitive Pressure Transducer

Enclosed in an airtight container is a metallic diaphragm which moves to the left when pressure is applied to the chamber and to the right when vacuum is applied. This diaphragm is used as one plate of a variable capacitor. Its distance from the stationary plate to its left, as determined by the pressure applied to the unit, determines the capacitance between the two plates. The monitor indicates the pressure equivalent of the unit's capacitance by measuring the capacitor's reactance to the ac source voltage.

(The portion of the chamber to the left of the moving plate is isolated from the side into which the pressurised gas or vapour is introduced. Hence, the dielectric constant of the unit does not change for different types of pressurised gas or vapour. The capacity is purely a function of the diaphragm position.) This device is not linear.

Changes in pressure may be easily detected by the variation of capacity between a fixed plate and another plate free to move as the pressure changes. The resulting variation follows the basic capacity formula.

$$C = 0.885\,\frac{K(n-1)A}{t}\ \text{pf} \tag{13.15}$$

where A = area of one side of one plate in cm^2

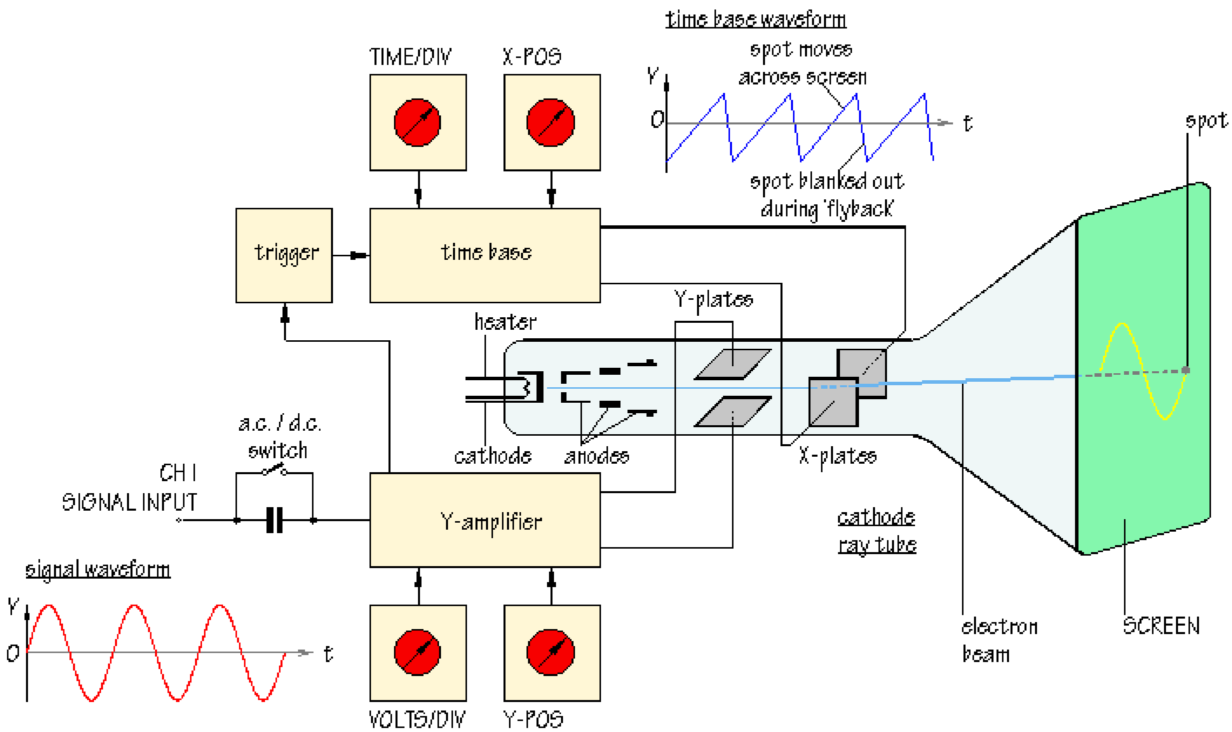
n = number of plates

t = thickness of dielectric in cm

CHAPTER 8- OSCILLOSCOPE

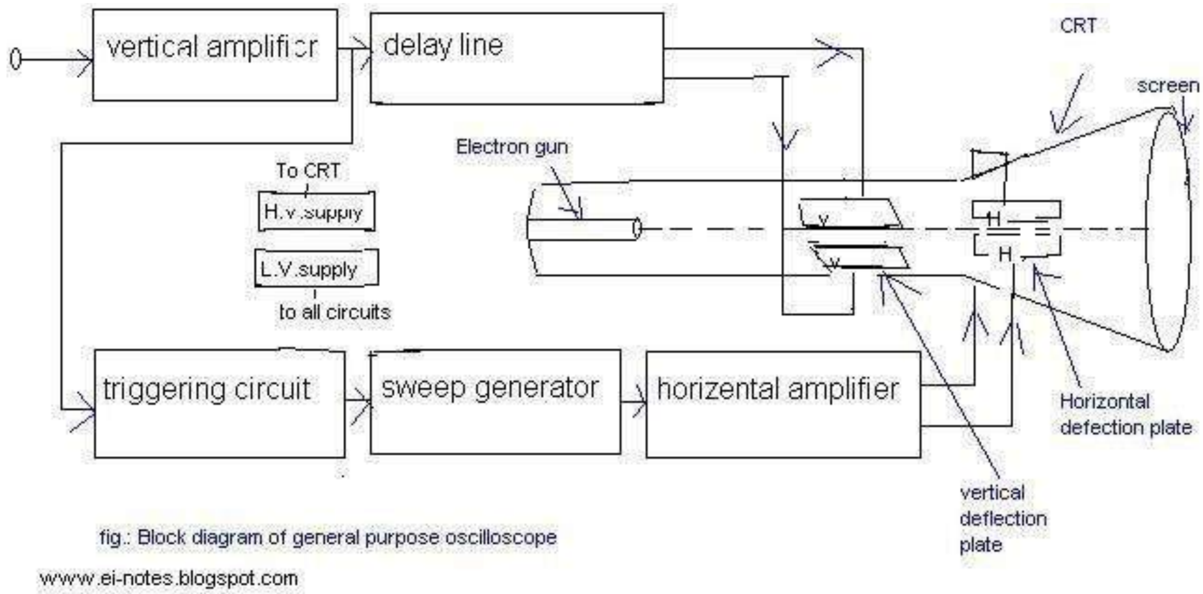
BASIC PRINCIPLE OF OSCILLOSCOPE.

A CRO (Cathode-Ray Oscilloscope), or DSO (Digital Storage Oscilloscope), is a typeof electronic test instrument that allows observation of constantly varyingsignalvoltages, usually as a two-dimensional plot of one or more signals as a functionof time.



BLOCK DIAGRAM OF OSCILLOSCOPE & SIMPLE CRO.

The block diagram of simple CRO is as shown in figure below.Herethe Oscilloscopes areused to observe the change of an electrical signal over time, such that voltage and timedescribe a shape which is continuously graphed against a calibrated scale. Theobserved waveform can be analyzed for such propertiesasamplitude, frequency,rise time,time interval, distortion and others. Modern digital instruments may calculate and displaythese properties directly. Originally, calculation of these values required manuallymeasuring the waveform against the scales built into the screen of the instrument.



The oscilloscope can be adjusted so that repetitive signals can be observed as a continuousshape on the screen. A storage oscilloscope allows single events to be captured by theinstrument and displayed for a relatively long time, allowing human observation of events toofast to be directly perceptible.Oscilloscopes are used in the sciences, medicine, engineering, and telecommunicationsindustry. General-purpose instruments are used for maintenance of electronic equipmentand laboratory work. Special-purpose oscilloscopes may be used for such purposes as analyzing an automotive ignition system or to display the waveform of the heartbeat asan electrocardiogram.

DUAL TRACE CRO:

The block diagram of dual trace oscilloscope which consist of following steps**,**

1. Electronics gun (single)
2.Separate vertical input channels (Two)
3. Attenuators
4.pr-amplifiers
5. Electronic switch.

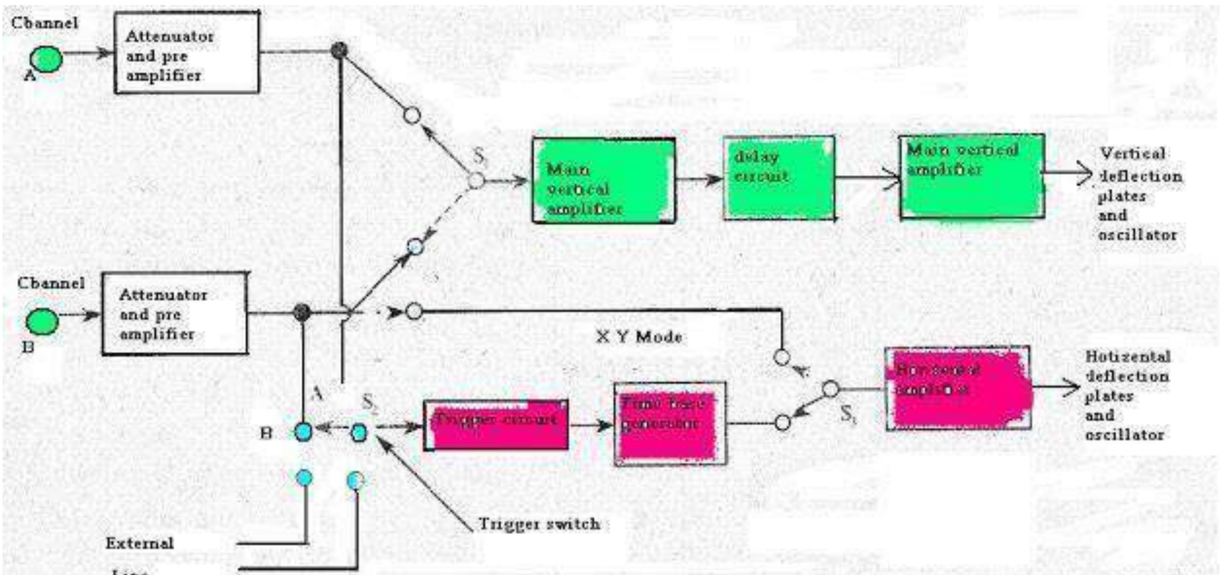


Fig: Block diagram of Dual Trace Oscilloscope

The two separate input signals can be applied to single electron gun with the helpof electronic switching it Produces a dual trace display .Each separate vertical inputchannel are uses separate attenuators and pr-amplifier stages, so the amplitude of eachsignal can be independentlycontrolled. Output of the pr-amplifiers is given to theelectronic switch, which passes one signal at a time into the main vertical amplifier of theoscilloscope.The time base-generator is similar to that of single input oscilloscope.By using switch S2 the circuit can be triggered on either A or Bchannel, waveforms, or an external signal, or on line frequency. The horizontalamplifier canbefed from sweep generator or from channel B by switching S1. When switch S, is in channelB, itsoscilloscope operates in the X-Y mode in which channel A acts as the vertical inputsignal andchannelBasthe horizontal inputsignal.

From the front panel several operating modes can be selected for display, like channel Bonly,channel A only, channels B and A as two traces, and signals A + B, A - B, B ~ A or - (A + B)as a single trace. Two types of common operating mode are there for the electronicswitch,namely,

1.Alternatemode

2.Chopmode**.**

GANDHI INSTITUTE OF EXCELLENT TECHNOCRATS BHUBANESWAR

DEPARTMENT OF ELECTRICAL ENGINEERING

LECTURE NOTES

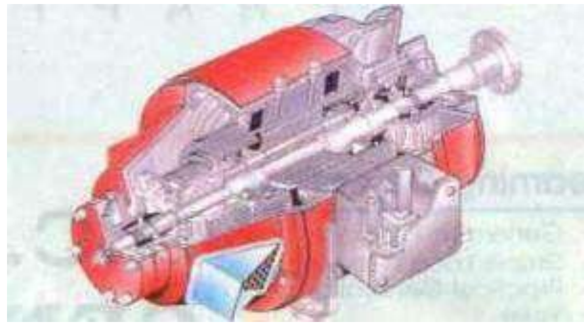
Year & Semester: 2ND Year IV Semester Electrical

Subject Code/Name: Th1. ENERGY CONVERSION

PREPARED BY:-

DC GENERATOR

An electrical generator is a machine which converts mechanical energy (or power) into electrical energy (or power).



The energy conversion is based on the principle Of the production Of dynamically (or motionally) induced e.m.f. As seen from fig. 26. I , whenever a conductor cuts magnetic flux. dynamically induced em-f. is produced in it according to Faraday's Laws of Electromagnetic Induction. This em-f. causes a current to now if the conductor circuit is closed.

Hence, two basic essential parts Of an electrical generator are a magnetic field and (ii) a conductor or conductors which can so move as to cut the flux.

CONSTRUCTION:-

In fig. 26. I is shown a single-turn rectangular coil ABCD rotating about its own axis in a

magnetic field provided by either permanent magnet is or electromagnets. The two ends of the coil

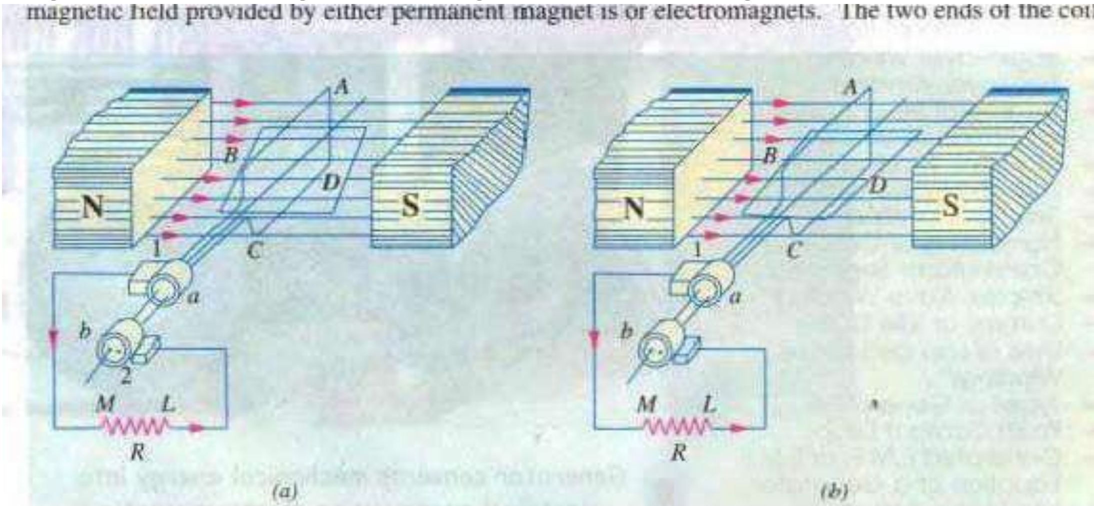


Fig. 26.1

magnetic field provided by either magnet is or electromagnets. The two ends of the coil

are joined to two slip-rings 'a' and •b' which are insulated from each Other and from the-central shaft. Two collecting brushes (of carbon or copper) press against the slip-rings. Their function is to collect the current induced in the coil and to convey il to the external load resistance R.

The rotating coil may bc called •armature' and the magnets as 'field magnets .

Working

Imagine the coil to be rotating in clock-wise direction (Fig. 26.2). As the coil assumes successive positions in the field. the flux linked with it Ehanges. Hence. an e.m.f. is induced in it which is

proportional to the rate Of change Of flux linkages (e = NdcÞdt). When the plane Of the coil is at right angles to lines of flux i.e. when it is in position, I, then flux linked with the coil is maximum butrate Ofchange Offlux linkages is minimum.

It is so because in this position, the coil sidesAB and CD do not cutor shear the flux, rather they slide along them they move parallel to them. Hence, there is no induced e.m.f. in the coil. Let us take this no-e.m.f. or vertical position of the coil as the starting position. The angle of rotation or time will be measured from this position.

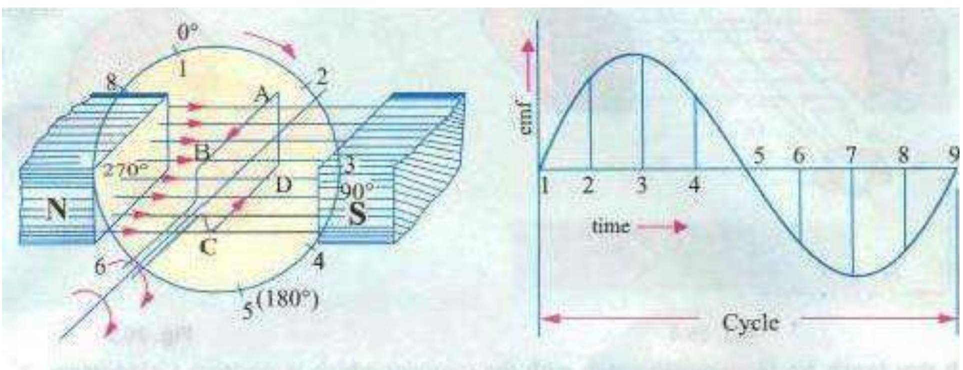


Fig. 26.2 Fig, 26.3

As the coil continues rotating further, the rate Of change Of flux linkages (and hence induced e_mf. in it) increases. till position 3 is reached where $6 = 90^0$. Here, the coil plane is horizontal i.e. parallel to the lines Of flux. As seen. the flux linked with the coil is minimum butrate ofchange Of 'lax linkages is maximum. Hence, maximum e.m.f. is induced in the coil when in this position (Fig. 26.3).

In the next quarter revolution i.e. from 900 to 1800, the flux linked With the coil gradually incrcríses but the rate Of change Of flux linkages decreases. Hence, the induced c.m.f. decreases gradually till in position 5 ofthe coil, it is reduced to zero value.

So, we find that in the first half revolution Of the coil, no (or minimum) e.m.f. is induced in it when in position l, maximum when in position 3 and noe.m.f. when in position 5. The direction Of this induced e.m.f. can be found by applying Fleming's Right-hand rule which gives its direction from A to B and C to D. Hence, the direction of current flow is ABMLCD (fig. 26. l). The current through the load resistance R flows from M to L during the first half revolution of the coil.

In the next half revolution i.e. from ISV to 360. the variations in the magnitude of are similar to those in the first half revolution. Its value is maximum when coil is in position 7 and minimum when in position I. But it will be found that the direction of the induced current is from D to C and B to A as shown in Fig. 26.1 Hence, the path Of current flow is along DUMBA which is just the reverse of the previous direction of now.

Therefore, we find that the current which we obtain from sueh a simple generator reverses its direction after every half revolution, Such a current undergoing periodic reversals is known as alternatingcurrent. It is. obviously, different from a direct current which continuously flows in one and the same direction. It should be noted that alternating current not only reverses its direction, it doesnoteven keep its magnitude constant while flowing in any one direction. The two half-cycles may be called positive and negative half-cycles respectively (Fig. 26.3).

For making the now Of current unidirectional in the external circuit. the slip-rings are replaced by split-rings (Fig. 26.4). The split-rings are made out of a conducting cylinder which is cut into two halves or segments insulated from each other by a thin sheet of mica or some other insulating material (Fig. 26.5).

As before. the coil endsare joined to these segments on Which rest thecarbon or brushes. It is seen IFig. 26,6 that in the first half revolution current flows along BMNI_CDJ i:e. the brush NO. I in contact with segment 'a' acts as the positive end of the supply and •b' the negative end. In the next half revolution IFig. 26.6 (b)l. the direction Of the induced current in the coil has

reversed.

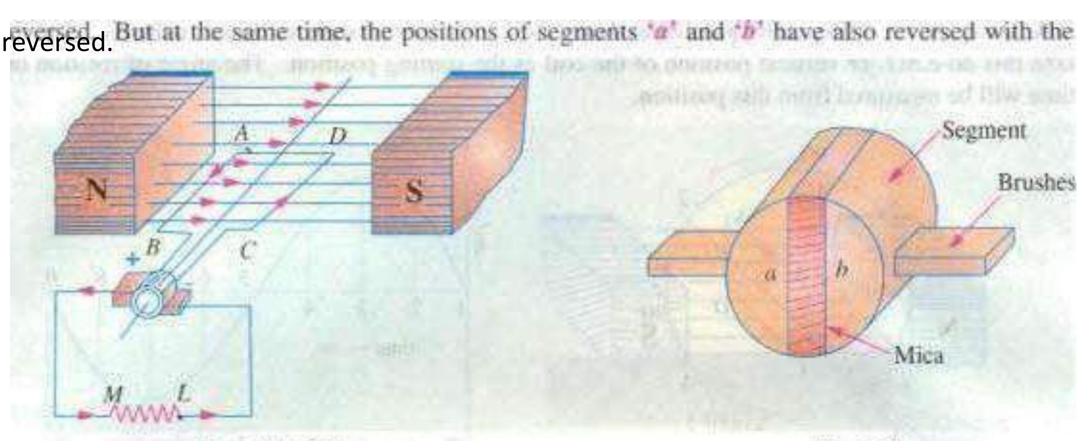
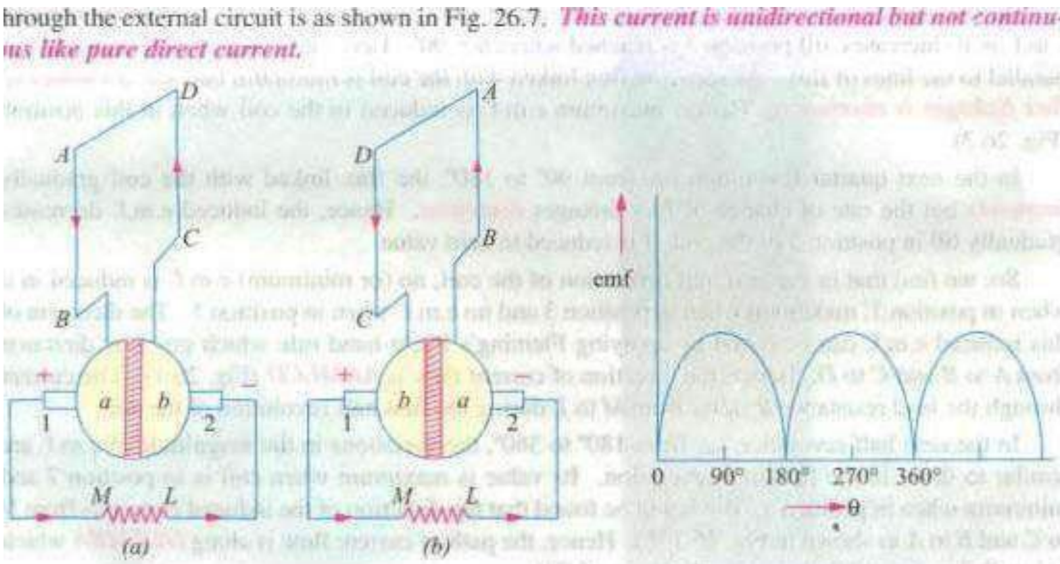


Fig. 26.4 Fig. 26.5 result that brush No. I comes in touch with the segment which is positive i.e. segment in this case. Hence. current in the load resistance again flows from M to



L. The Waveform of the current through the external circuit is as shown in Fig. 26.7. This current should be noted that the position ofbrushes is so arranged that the change over Of segments •a' and from one brush to the other takes place when the plane of the rotating coil is at right angles to the plane of the lines of It is so because in that position. the induced e.m.f. in the coil is zero.

Another important point worth remembering is that even now the current induced in the coil sides is alternating as before. It is only clue to the rectifying action of the split-rings (also called commutator) that it becomes unidirectional in the external Circuit. Hence. it should be clearly undEr• stood that in the armature Of a generator. the induced voltage is alternating.

Generator-

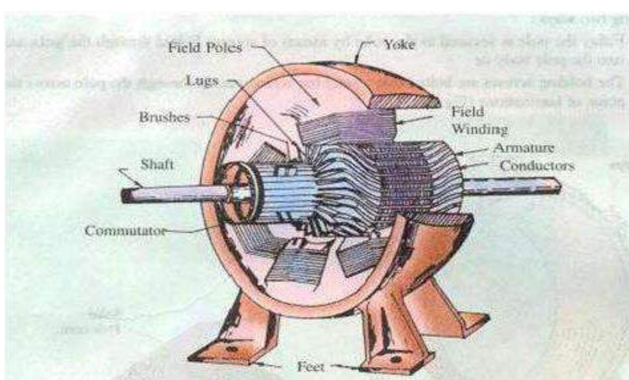


Fig. 26.8

Pole Cores and Pole Shoes

The field Consist or pole Cores and pole shoes. The pole shoes SCñre two purposes

(i) they spread out the flux in the air gap and also, Of larger cross-section, reduce the reluctance Of the magnetic path • ii) they supiX3rt the exciting coils (or field coils) as shown in fig. 26.14.

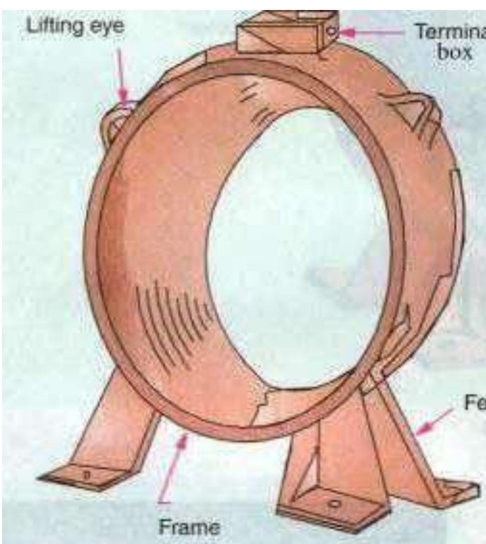
There are two main types Of pole construction.

(a) The pole core itself may a solid piece made out Of eithercast iron Or cast steel but the pole shoe is laminated and is fastened to the pole face by means of counter sunk screws as shown in fig. 24.10.

(b) In modern design. the complete Ne cores and pole shoes are built of thin laminations of annealed steel which aœ rivetted together under hydraulic pressure (Fig. 26_11). The thickness Of laminations varies from I mm to 0.25 mm- The laminated poles may be secured to the yoke in any of the following two Ways:

(i) Either the pole is secured to the yoke by means of screws bolted through the yoke and into the pole body or

(ii) The holding screws are bolted into a bar which passes through the pole across the plane of laminations (Fig. 26.12).



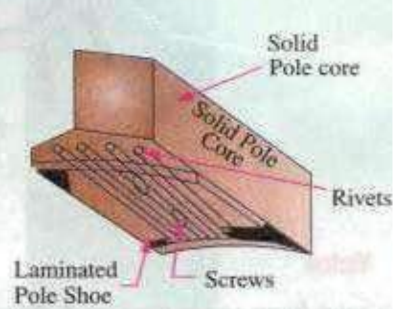
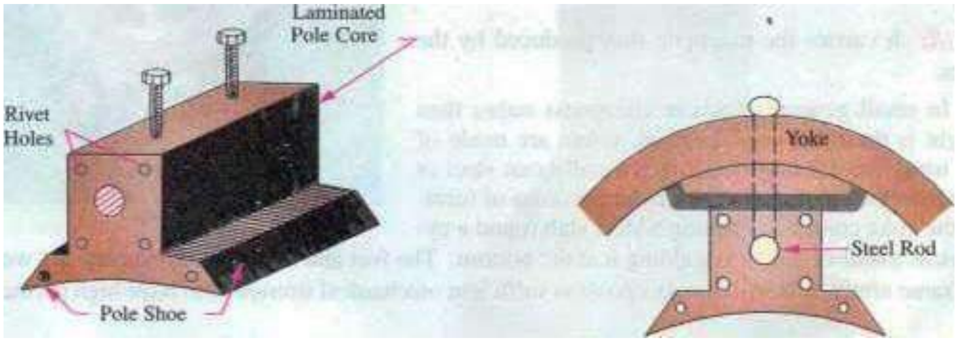


Fig. 26,9 Fig. 26.10



pole coas

The field coils or pole coils, which consist of copper wire or strip. are former-wound for the correct dimension (Fig. 26.13). Then. the former is removed and wound coil is put into place over the core as shown in Fig. 26.14.

When current is passed through these coils, they electromagnetise the poles which produce the necessary flux that is cut by revolving armature conductors.

Armature Core

It houses the armature conductors or coils and causes them to rotate and hence cut the magnetic flux Of the field magnets In addition to this, its most important function is to provide a path of very low reluctance to the flux through the armature from a tv-pole to a S-pole.

It is cylindrical or drum-shiped and is built up of usually circular sheet steel discsorlaminations approximately 0.5 mm thick (Fig. 26.15). It is keyed to the shaft.

The slots are either die-cut or punched on the outer periphery Of the disc and the keyway is located on the inner iameter asshown. In small machines, the armature stampings are keyed directly to the shaft. Usually. these laminations are perforated for air ducts which permits axial now of air through the armature for cooling purposes. Such ventilating channels are clearly Visible in the laminations shown in Fig. 26.16 and Fig. 26.17.

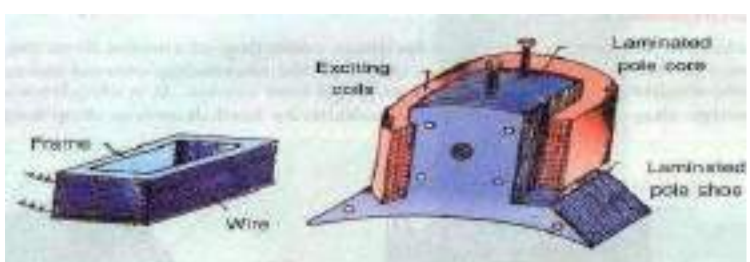
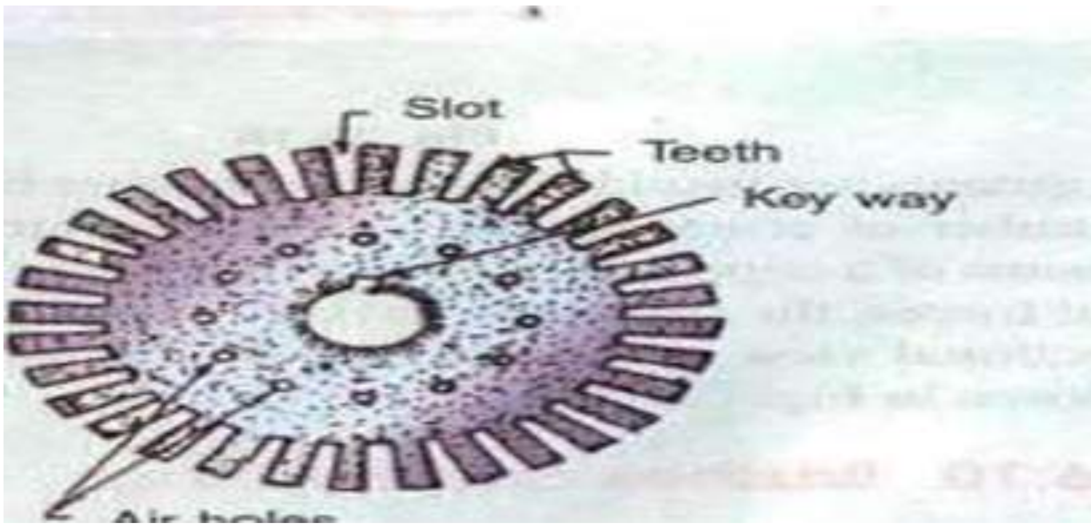
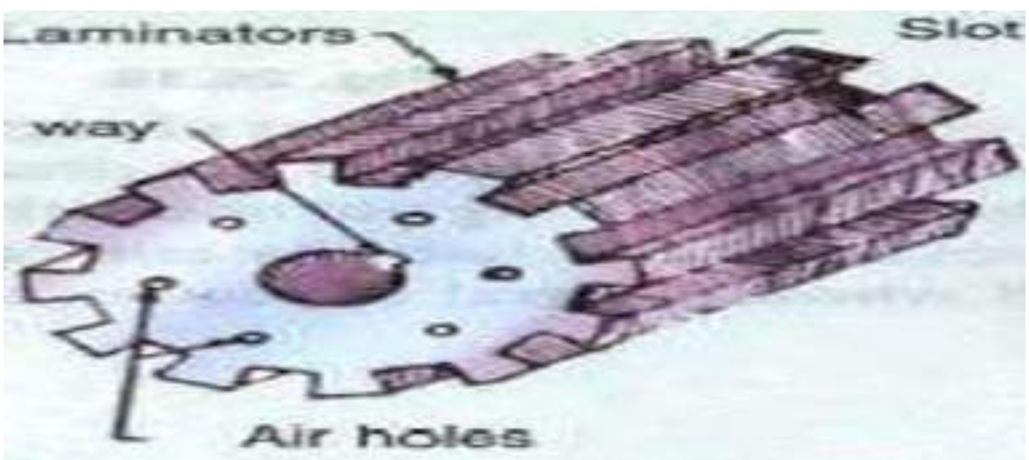


Fig. 26.13 Fig. 26.14

Up to armature diameters Of about one metre, the circular stampings are cut out in one piece as shown in Fig. 26.16, But above this size. these circles, especially of such thin sections, are difficult to handle because they tend to distort and become wavy when assembled together. Hence. the circular laminations. instead of being cut out in one piece, are cut in number Of suitable sections or segments which form part ofa complete ring (Fig. 26.17).





A complete circular lamination is made up Of four or six or even eight segmental laminations. Usually. two keyways are notched in each segment ana are dove-tailed or wedge-shaped to make the laminations self-locking in position,

The purpose of using laminations is to reduce the loss due to eddy currents. Thinner the laminations. greater is the resistance offered 10 the induced e_rn.f., smaller the current and hence lesser the R loss in the core.

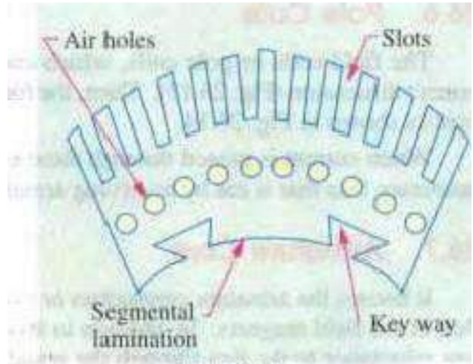


Fig. 26.17

s and are then pulled into their proper shape

Armature Windings

The armature windings are usually former-wound.

These are first wound in the form of nat rectangular coils and are then pulled into their in a coil puller. Various conductors Of the coils are insulated from each other. The conductors are placed in the armature slots which are lined With tough insulating material. This Slot insulation is folded over above the armature conductors placed in the slot and is secured in place by special hard wooden or fibre wedges,

Commutator

The function ofthe is to facilitate collection of current from the armature conduc• torS. As shown in Art. 26.2, rectifiedie•. converts the alternating current induced in the armature conductors into unidirectional current in the external load circuit. It is of cylindrical structure and is built up of wedge-shaped segrtients of high•conductivity hard-drawn or drop forged copper. These

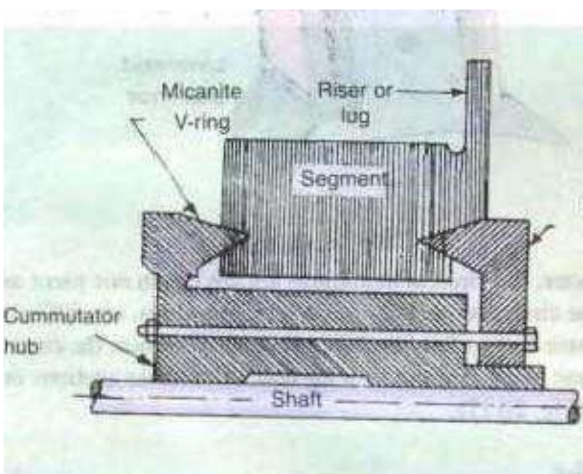


Fig. 26.18

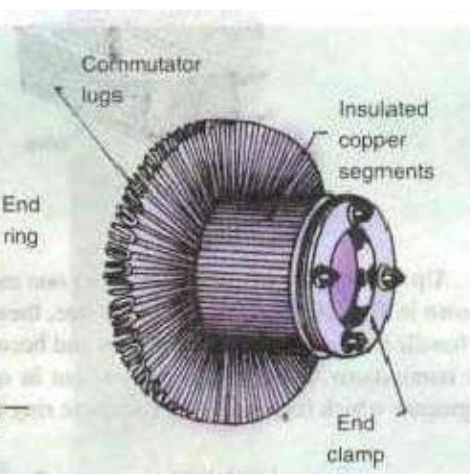


Fig. 26.19

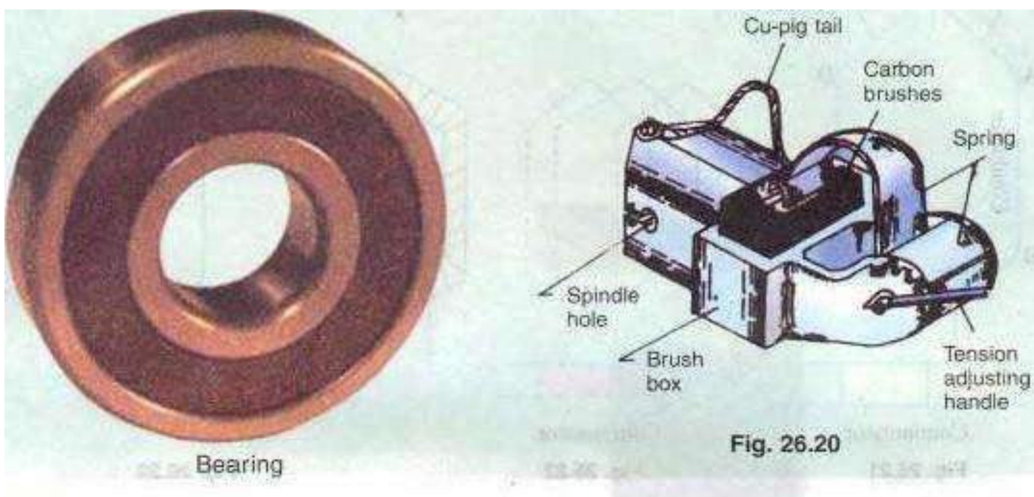
segments areinsulated from each Other by thin layers Oftliiea. [L] The number Of segments is equal to the number of armature-coils. F.qch commutator segment is connecled to the armature conductor by means of a copperÌugoi strip (or riseth To prevent them from flying out under the action of centrifugal forces, the Segments have V-grooves. these grooves being insulated by conical micanite rings. A sectional view Of commutator is in fig. 26. IS Whose general When completed is shown in Fig. 26.19.

Armature winding

Brushes and •Bearings

The brushes function is to collect current from commutator.are usually made Of carbon or graphite and are in the shape of a rectangular block. These brushes are housed in brush-holders usually of the box-type variety. As shown in Fig. 26.20. the brush-holder is mounted a spindle and the brushesean slide in the rectangular-box Open at both ends. The bruéhes madeto bear down On the Commutator by a spring whose tension Cán be adjusted by Changing the 1Sositioñ Of lever in the notches. A flexible copper pigtail mounted at the top of the brush conveys current from the brushes to the holder. The number of brushes per spindle depends on the magnitude Of the corrcnt to be collected from the commutator_

Flexible



Bearing

Fig. 26.20

Because Oftheirreliahility, hall-bearings are frequently eniployed. though for heavýdutieS, roller bearingsare preferable. The ball and rollers are generally packed in hard oil for quieter operation and for reduced bearing wear, sleeve bearingé are used which are lubricated by ring oilers fed from Oil reservoir in the bearing bracket.

Armature Windings

NOW, We Will discuss the winding Of an actual armature. But before doing this, the meaning Of the following terms used in connection With armature winding should be clearly kept in mind, Pole-pitch

It may be variously defined as :

(i) The periphery Of the armature divided by the number of poles orthe generator i.e. the distance between two adjacent poles.

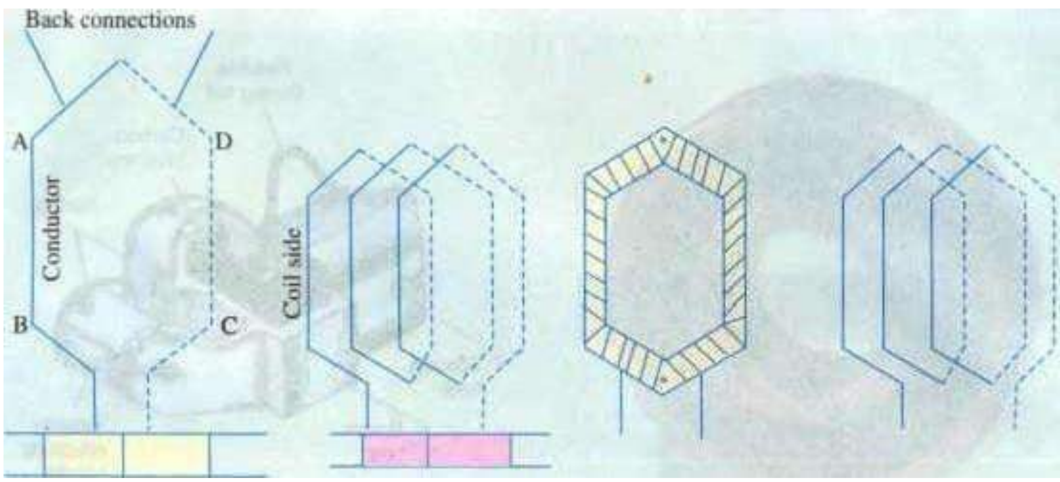
(ii) It is equal to the number Of armature

•conductors and 4 poles, the pole pitehis 48/4 12.

Conductor

The length Of a Wire lying in the magnetic field and in which an e.mf. is induced, is calleda conductor (or iyduçtot) for example, length AB orCD in Fig. 26_21_

With reference to Fig. 26.21 , the two conductors AB and CD along With their end connections constitute one coil Of the armature winding. The coil may be single-turn coil (Fig. 26.21) or multiturn coil (Fig. 26.22). A single-turn coil will have two conductors. But a multi-turn coil may have many conductors per coil side, In Fig, 2622, for example. each coil side has 3 conductors, The



Cornmutator Commutator

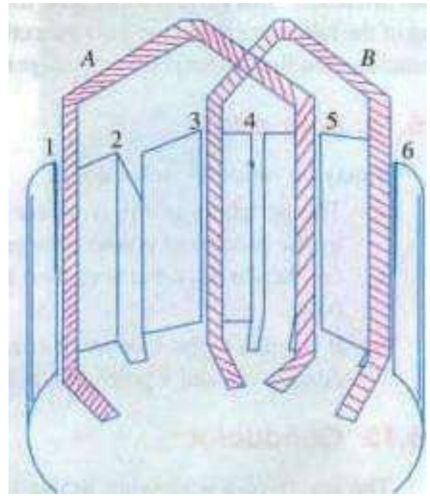
Fig. 26.21 Fig. 26.22 Fig. 26.23

group of wires or conductors constituting a coil side of a multi-turn coil is wrapped with a tape as a unit (Fig. 26.23) and is placed in the armature slot. It may noted that since the beginning and the end Of each coil must be connected to a commutator bar. there are as many commutaror bars as coils for both the lap and wave windings (see Example 26. I).

The side of a coil (I-turn or multiturn) is.illed a winding element. Obviously, the number of winding elements is twice the number of coils.

Col-spm or Coil-pitch CYs)

It is the distance, measured in terms of armature slots for armature conductors) between two sides of a coil. It is, in fact, the periphery of the armature spanned by the two sides of the coil.



If the pole span or coil pitch is equal to the pole pitch (as in the case of coil A in Fig. 26.24 where polepitch of 4 has been assumed). then winding is called fall-pitched. It means that coil span is ISO electrical degrees. In this case. the coil sides lie under opposite poles, hence the induced e.rn.fs. in them are additive. Therefore, maximum e.m.f. is induced in the coil as a Whole. it being the sum Of the e.m.f.s induced in the two coil sides. For example. if there are 36 slots and 4 poles. then coil span is 36/4 = 9 slots. If number of slots is 35, then Ys = 35/4 = 8 because it is customary to drop

If the coil span is less than the pitch (as in Coil Fig. 26.24

B where coil pitch is 314th of the pole pitch), then the

winding is fractional-pitchea_L In this case, there is a phase difference sides of the
Hence, the total e_m.f. round the coil Which is the vector sum of e.m.fs. inAhe two coil sides, is less in this caseas compared to that in the first case.

Pitch of a Winding (Y)

In general, it may be defined as the distance round the armature between two successive conductors which directly connected together. Or, it is the distance between the beginnings Of two consecutive turns.

$-Y_F$ —for lap winding

$+Y_F$—-„.for wave winding

In practice. coil-pitches as low as eight-tenths of a pole pitch are employed without much serious reduction in the e.m.f. Fractional-pitched windings are purposelyused to effect substantialsaving in the copper Of thecnd connections and for improving commutation.

Back Pitch

The distance, measured in terms of the armature conductors, which a coil advances on the back of the armature is called back pitch and is denoted by YR

As scen from fig_ 26_28, element I is connected on the back Of the armature to element 8. Hence. $Y_B = (8 - 1) =$ 7.

Front Pitch (Yr)

The number of armature conductors or elements spanned by a coil on the front (or commutator end of an armature' is called the front pitch and is designated by Again in Fig. 26.28. element 8 is connected to elemcnt 3 on the front Of the armature, the connections made at the commutator segment. Hence, YF = 8 — 3 = 5.

Alternatively. the front pitch may be defined as the distance (in terms Of armature conductors) between the second conductor of one coil and the first conductor of the next coil which are connected together at the front i.e. commutator end of the annature. Both front and back pitches for lap and wave-winding are shown in fig. 26.25 and 26_26.

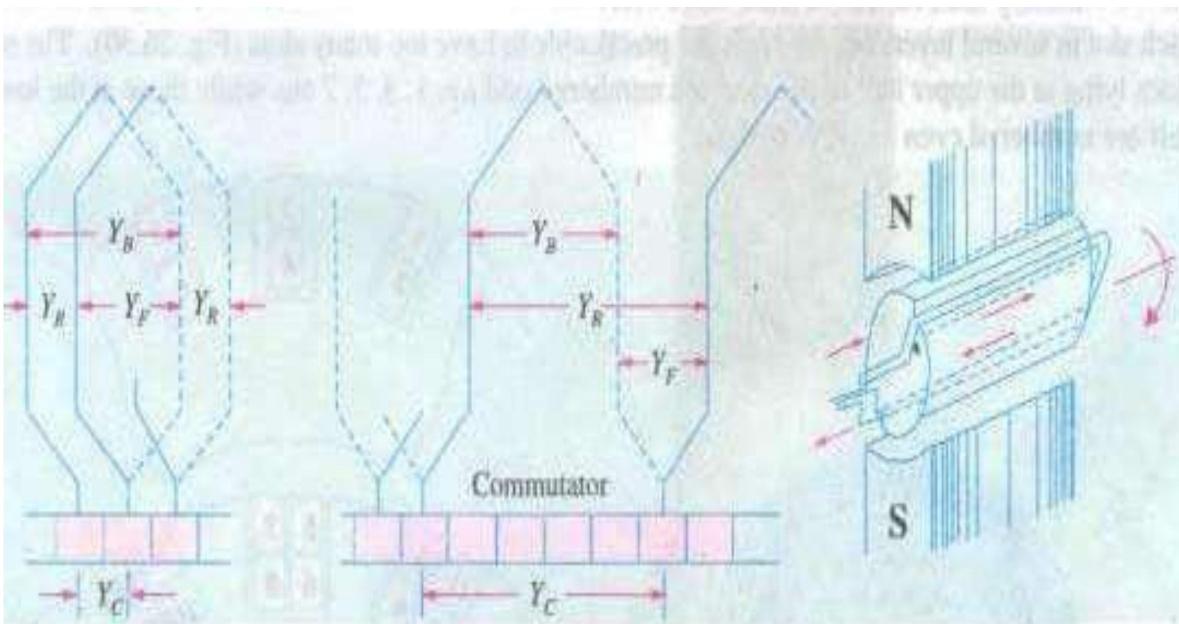


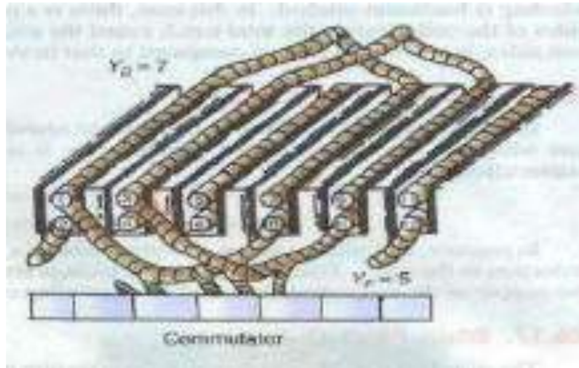
Fig. 26.25 Fig. 26.26 Fig. 26.27

Resultant Pitch

It is the distance between the beginning of one coil and the beginning of the next coil to which it is connected (Fig. 2625 and 26.26).

As a matter of precaution. it should be kept in mind that all these pitches. though normally stated in terms Of armature conductors. are also sometimes given in terms Of armature slots or commutator bars because commutator is, after all, an image Of the winding.

Cornmutator Pitch (Y



It is the distance (measured in commutator bars or segments) between the segments to which the two ends of a coil are connected. From Fig. 26.25 and 26.26 it is clear that for lap winding. Yc is the difference of YB and Y_F whereas for wavewinding it is the sum of Y_B and Yr Obviously, commutator pitch is equal to the number of bars between coil leads. In general, equals the 'plex• of the lap-wound armature. Hence, it is equal l, 2, 3, 4 etc. for simplex-. duplex, triplex—and quadruplex etc. lap-windings. Fig. 26.28

26-21. Single-layer Winding

It is that winding in which one conductoror one coil side is placed in each armature slot as shown in Fig. 26.27. Such a winding is not much used.

Two-layer Winding

In this of winding, there are two conductors or coil sides Ixr slot arranged in two layers. Usually, one side of every coil lies in the upper half of one slot and other side lies in the lower half of some other slot al a distance of approximately one pitch away (Fig. 26.28). The transfer of the coil from one slot to another is usually made in a radial plane by means of a peculiar bend or twist at the back end as shown in Fig, 26.29. Such windings in which two coil sides occupy each slot are most commonly used for all medium-sized machines. Sometimes 4 Or 6 or 8 coil sides are used in each slot in several layers because it is not practicable to have too many slots (fig. 26.30). The coil sides lying at the upper half of the slots are numbered odd i.e. I , 3.5, 7 etc. while those at the lower half are numbered even i.e. 2, 4. 6, 8 etc.

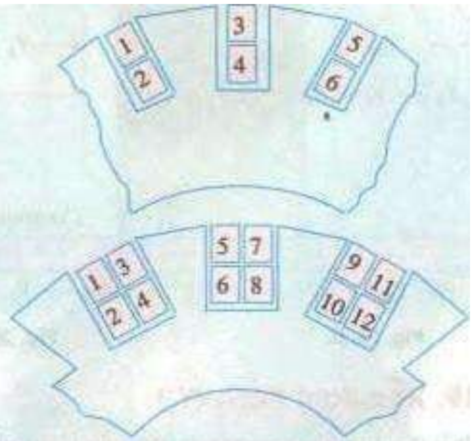


Fig. 26.29   Fig. 26.30

Degree of Re-entrant Of an Armature WindingA

winding is Saidto be Single re-entrant if on tracing through it once, all armature conductors are included on returning to the starting point. It is double re-entrant if only half the conductors are included in tracing through the winding once and so On,

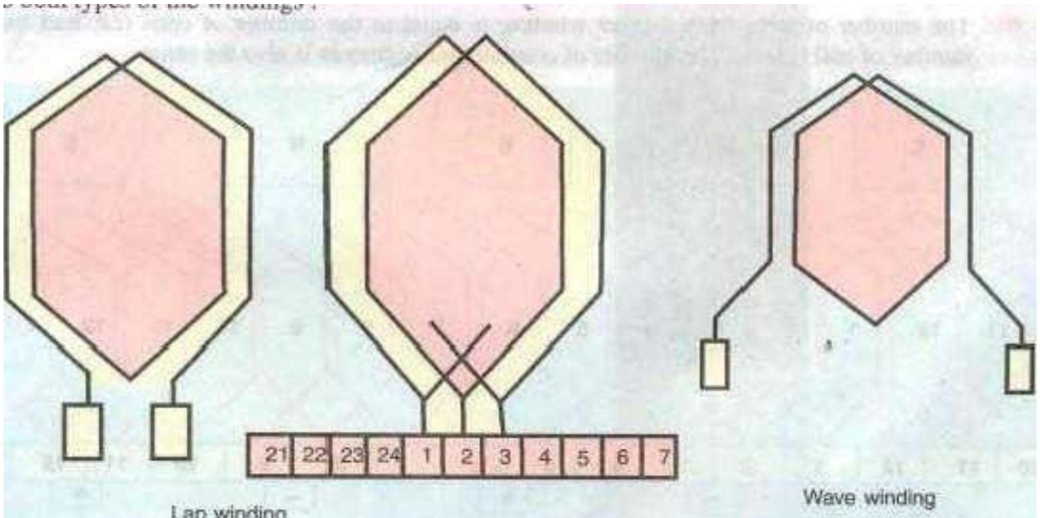
Multiplex Winding

Multiplex Winding

Insuch windings. there are several sets ofcompletely closed and independent windings. If there is only one set of closed winding, it is called simplex wave winding. If there are two such windings on the same armature, it is called duplex winding and so on. The multiplicity affects a number of parallel paths in the armature. For a given number of armature slots and coils, as the multiplicity increases, the number Of parallel paths in the armature increases thereby increasing the current rating but decreasing the voltage rating.

Lap and Wave Windings

Two types ofwindings mostly employed for drum-type armatures are known as Lap Winding and Wave Winding. The difference between the two is merely due to the dtfferentarrangement of the end connections at the front or ommutatorend of armature.



Lapwinding

Each winding can be arranged progressively or retrogressively and connected in simplex, duplex and triplex. The following rules, however, apply to both types of the windings :

(i) The front pitch and back pitch are each approximately equal to the pole-pitch i.e. windings should be full-pitched. This results in increased e.m.f, round the coils. For pur-

poses. fractional-pitched windings are deliberately used (Art_ 26.15).

(in Both pitches should be odd, otherwise it would be difficult to place the coils (which are former-wound) properly on the armature. Forexmaple, ifYB and YFwere both even, the all the coil Sides and conductors would lie either in the upper half Of the slots Or in the lower

half. Hence. it would become impossiblefor one Aidèxsfthe coil in the upper half. Hence. it would become impossible One side of the coil to lie in the upper half of one slot and the other side of the same coil to lie in the lower half of other slot.

riii', The of commutator segmentsis equal to the number of slots or coils (or half the number Of conductors) because the front endsÃ'f conductors are joined to the segments in

The winding mustclose uponitselfi.e. if we start frotnû given point and move from onecoi] to•another, then all conductors should he traversed and we should reach the same poinragain Without a break or discontinuity io between.

Simplex Lap-winding*

It is shown in fig: 26-25 which employs single-turn coils. In lapwinding, the finishing end of one coil is connected to a commutator segment and to the starting end of the adjacent coil situated underthe same pole and so on, till and the coils have been connected. This type Of winding derives its name from the fact it doubles or laps hack With its succeeding cods.

Following points regarding simplex lap winding should be

1. The back and front pitches are •odd•and Of opposite Sign. But they Cannot They differ by 2 Or Some multiple thereof.

2. Both and should be nearly equal toa pole pitch.

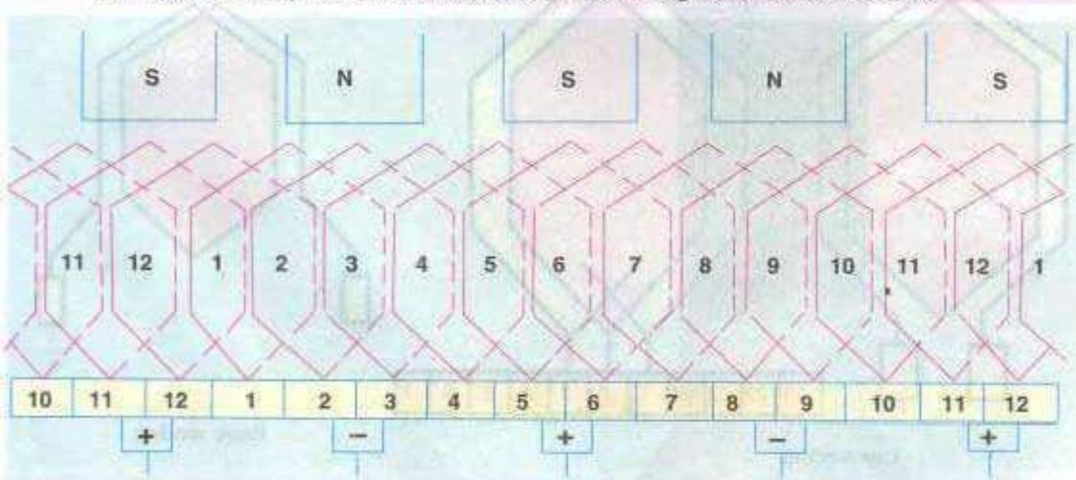
3. The average pitch YA = . It equgls pole pitch = z

4. pitch = ± I (In keneral, = ± m)

S. Resultant pitch YR is even, being the arithmetical difference of two odd numbers, i.e.,

6. The number of slots for a 2-layer winding is equal to the number bf coils (i.e. half the

number of coil sides). The number of commutator segments is also the same.



Simplex lap winding

* However, where heavy currents are necessary, duplex or triplex lap windings are used. The duplex lap winding is obtained by placing two similar windings on the same armature and connecting the even-numbered commutatorbarstoone winding and 'heodd-numberedones tothesecond winding. Similarly.

in triples Winding. there would three Windings, cach connected to third of the commututor bars,

7. The numberofpárallel pathsin the armature where m is the multiplicityofthe winding and P the number of poles.

Taking the firstcondition. we have Yo = $Y_F \pm 2$.

(a) If YB > YF i.e. = YF • 2, then We get a progressive or right-handed winding i.e. a winding which progresses in the clockwise direction as seen from the commutator end.o this case, obviously,

$Y_C = +1$.

(b) IIf YB < YF i.e. YF —2, then we geta or left-handed winding i.e. one which advances in the anti-clockwise direction When seen from the commutator side. In this Case, Yc+-1.

(c) Hence, it is obvious that

$$\left.\begin{aligned} Y_F &= \frac{Z}{P} - 1 \\ Y_B &= \frac{Z}{P} + 1 \end{aligned}\right] \text{for progressive}$$

SQinding ûild

$$\left.\begin{aligned} Y_F &= \frac{Z}{P} + 1 \\ Y_B &= \frac{Z}{P} - 1 \end{aligned}\right]$$

— + 1

forretrogfesgive winding

Obviously, up must be even to make the winding possible.

Numbering Of Coils and Commutator Segments

In the d.c. Winding diagrams to follow. we Will number the coils only (not individual turns). The upper side Of the coil Will be shown by a firm continuous I inc whereas the lower side Will be shown by a broken line. The numbering of coil sides will be consecutive i.e. l, 2, 3 etc. and such that odd numbers are assigned to the top conductors and even numbers to the lower sides for a two-layer winding. The commutator segments Will also be numbered consecutively, the of the segments will be the same as that of the upper side connected to it.

Example 26.1. Draw a developed diagram ofa simple 2-laýer lap-windingfor "-pole generator With 16 coilÅ. Hence, point out the chamcteristics Of a lap-winding.

(Elect. Engineering, Madras Univ. 1981)

Solution. The number of commutator segments = 16

Number Of conductors or coil sides 16 x 2 32 : pole pitch 32/4 = 8

Now remembering that (i) YB and YF have to be Odd and (ii) have to differ by 2, We get for a progressive winding YB = 9 ; YF — —7 (retrogressive winding Will result if YR 7 and YF = -9). Obviously, commutator pitch Yc — -1.

(Otherwise, as shown in An. 26.26, for progressive winding

$Y_F = \frac{Z}{P} - 1 = \frac{32}{4} - 1 = 7$ and $Y_B = \frac{Z}{P} - 1 = \frac{32}{4} + 1 = 9]$

The Simple winding table is given as under:

Beck Connections		Front Connections
1 to (1 + 9) = 10	$\longrightarrow$	10 to (10 − 7) = 3
3 to (3 + 9) = 12	$\longrightarrow$	12 to (12 − 7) = 5
5 to (5 + 9) = 14	$\longrightarrow$	14 to (14 − 7) = 7
7 to (7 + 9) = 16	$\longrightarrow$	16 to (16 − 7) = 9
9 to (9 + 9) = 18	$\longrightarrow$	18 to (18 − 7) = 11
11 to (11 + 9) = 20	$\longrightarrow$	20 to (20 − 7) = 13
13 to (13 + 9) = 22	$\longrightarrow$	22 to (22 − 7) = 15
15 to (15 + 9) = 24	$\longrightarrow$	24 to (24 − 7) = 17
17 to (17 + 9) = 26	$\longrightarrow$	26 to (26 − 7) = 19
19 to (19 + 9) = 28	$\longrightarrow$	28 to (28 − 7) = 21
21 to (21 + 9) = 30	$\longrightarrow$	30 to (20 − 7) = 23
23 to (23 + 9) = 32	$\longrightarrow$	32 to (32 − 7) = 25
25 to (25 + 9) = 34 = (34 − 32) = 2	$\longrightarrow$	2 to (34 − 7) = 27
27 to (27 + 9) = 36 = (36 − 32) = 4	$\longrightarrow$	4 to (36 − 7) = 29
29 to (29 + 9) = 38 = (38 − 32) = 6	$\longrightarrow$	6 to (38 − 7) = 31
31 to (31 + 9) = 40 = (40 − 32) = 8	$\longrightarrow$	8 to (40 − 7) = 33 = (33- 32) = 1

** In general, $Y_B = Y_F \pm 2m$ where $m = 1$ for simplex lap winding and $m = 2$ for duplex lap winding etc.

The winding ends here because we come back to the conductor from where we started.

We will now discuss the developed diagram which is one that is obtained by imagining the armature surface to be removed and then laid out flat so that the slots and conductors can be viewed without the necessity of turning round the armature in order to trace out the armature windings. Such a diagram is shown in Fig. 26.31.

Front end of the upper side of coil No. 1 is connected to a commutatorsegment (whose number is also I The backend is joined at the back to the I + 10th coil side in the lower half of5th slot. The front end Of coil side 10 is joined to commutator segment 2 to which is connected the front end of 10 —7 3 i.e. 3rd coil side s N lying in the upper half of second armature slot. In this way. by 26.32 travelling 9 coil sides to the right at the back and 7 to the left at the

front we complete the winding, thus including every coil side once till we reach the coil side I from where we started. Incidentally, it should be noted that all upper coil sides have been given Odd numbers, whereas lower ones have been given even numbers as shown in the polar diagram (Fig. 26.32) Of the Winding of Fig. 26.31.

Brush positions can be located by finding the directionofcurrents flowing in the various conductors. Ifcurrents in the conductors under the influence of a N-pole are assumed to flow downwards (as shown), then these will now upwards in conductors under the influence of S.pole. By putting proper arrows on the conductors (shown separately in the equivalent ring diagram), il is found that commutator bars NO. I and 9 are the meeting points Of c.m_fs. and hence currents are flowing out Of these conductors. The positive brushes should, therefore. be placed at these commutator bars. Similarly, commutator bars No. 5 and 13 are the separating points of e.m.fs. hence negative brushes are placed there. In all. there are four brushes, two positive and two negative. If brushes of the same polarity are connected together, then all the armature conductors are divided iuto four parallel paths.

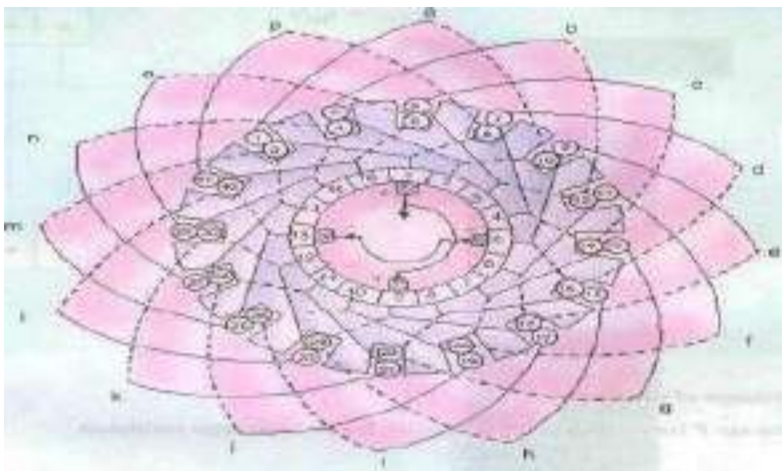


Fig. 26.33

Division of conductors into parallel paths is shown separately in the schematic diagram of Fig.

26.34. Obvi0LLsIy, if is the total current supplied by the Eenerator, then current carried by each parallel path is

Summarizing these conclusions, we have

1. The total number of brushes is equal to the number of poles.

2. Thereareas many parallel paths in the armature as the numberofpoles. That is why such a winding is sometimes as 'multiple circuit' or •parallel' winding. general, number Of parallel paths in armature mp where m is the multiplicity (plex) Of the lap winding. For example, a duplex lap winding has (6 x 2) = 12 parallel paths in its armature.

3. The e.m.f. between the +veand—ve brushesisequa/ to the e.m.f- generated in any one of the parallel paths. If Z is the total number of armature conductors

and p the number of poles, then the number of armature conductors (connected in series) in any parallel path is Z"

Resistance of each path— PIZ

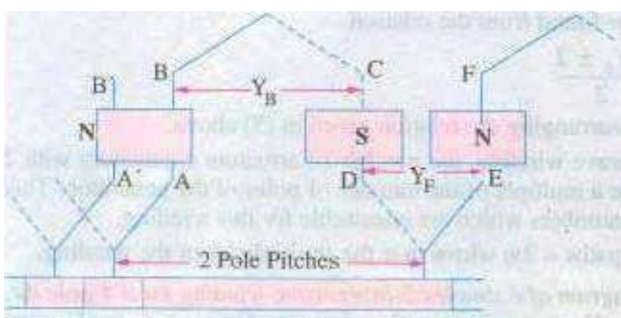
There are P (or A) such paths in parallel, hence equivalent resistance

5 If is the total armature current. then current per parallel path (or carried by each conductor) is a/P.

Simplex Wave Winding•

From fig. 26.31. it is clear that in lap Winding, a Conductor (or coil side) under one pole is connected at the back to a conductor which occupies an almost corresponding position under the next pole of opposite' polarity (as conductors 3 and 12). Conductor NO. 12 is then connected to conductor No. 5 under the original pole but which is a little removed from the initial conductor No. 3. If, instead of returning to the same At-pole, the conductor No. 12 were takenfonvard lothe next N.pnle, it would make no difference so fartLs the direction and magnitude of the e.m_f. induced in the circuit are concerned

Like lap winding. a wave winding may be duplex, triplcž or may have any degree Of multiplicity. A simplex wave winding has two paths. a duplex A',eave Winding four paths and a triplex paths etc.



As shown in Fig. 26.35. conductor AB Lis connected to CD lying under S-pole and then to EF under the next N•p01e, In this way, the winding, progresses, passing successively under every N-pole and S-pole till it returns to a conductor A 'B' lying under the original pole. Because the winding progresses in one direction round the armature in series of •waves' , it is known as wave Winding.

Fig. 26.35 If. after passing once round the armature. the winding falls in a slot to the left of its starting point A'B' in Fig. 26.35) then the winding is said to bc retrogressive. If, however, it falls one slot to the right. then it is progressive.

Assuming a 2-layer winding and supposing that conductor AB lies in the upper half of the slot. then going once round the armature, the Winding ends at Which most be atThe upper half of the slot at the left orTight. Counting in terms of conductors, it means that AB and A'B' differ by two conductors (although they diner by one slot).

From the above, we can deduce the following relations. If P —No. Ofpoles, then total No. Of conductors orcoil Sides

$$Y_B = \text{back pitch}, \quad Y_F = \text{front pitch} \quad \Big\} \text{ nearly equal to polepitch}$$

$$Y_A = \frac{Y_B + Y_F}{2} = \text{average pitch} ; Z =$$

Since p is always even and

$$Y_A \times P = Z \pm 2 \qquad Y_A = \frac{Z \pm 2}{P}$$

…s even and Z = PYR ± 2, hence Z must alwaysbeeven. Put in another way. it

means that Z must an even integer.

The plus sign will give a progressive winding and the negative sign a retrogressive winding.

Points to Note :

1. Both pitches YB and YFare odd and of the same sign.

2. Back and front pitches are nearly equal to the pole pitch and unay be equal or differ by 2, in which case. they are respectively one more or one less than the average40itch. 3. Resultant pitch YR $= Y_F + Y_B$.

4. Commutator pitch, $Y_C = Y_A$ (in lap winding $Y_C = \pm 1$).

Also.

$$Y_C = \frac{\text{No. of Commutator bars} \pm 1}{\text{No. of pair of poles}}$$

5. The average pitch which must be an integer is given by

$$Y_A = \frac{Z \pm 2}{P} = \frac{\frac{Z}{2} \pm 1}{P/2} = \frac{\text{No. of Commutator bars} \pm 1}{\text{No. of pair of poles}}$$

which It is clear that 10 he an integer. there is a restriction on the.alue of Z With Z = 32. this winding i'. impossible for 4-polc machine (though lap winding is possible). Valuesor Z = 30 or 34 would be perfectly ah-ight_

i.a NC can be found from the relation, PYA ±2

This relation has been found by rearranging the relation given in (5) above.

7. rt is obvious from (5) that for a wave winding, the number of armature conductors with 2 either added or subtracted must be a multiple of the number of poles of the generator. This restriction eliminates many even numbers which are unsuitable for this winding.

S. Thé number of armature parallel paths = 2m where m is the multiplicity of the winding.

Example 26.2. Draw a developeddiagram Ofa simplei -layer wave-winding for a 4-pole dc. generator With 30 armature conductors. Henée, point Out the characteristics Of a simple wave winding. (Elect. Engg-I, Nagpur Univ. 1991)

Solution. Here, $Y_A = \dfrac{30 \pm 2}{4} = 8^*$ or 7. Taking $Y_A = 7$, we have $Y_B = Y_F = 7$

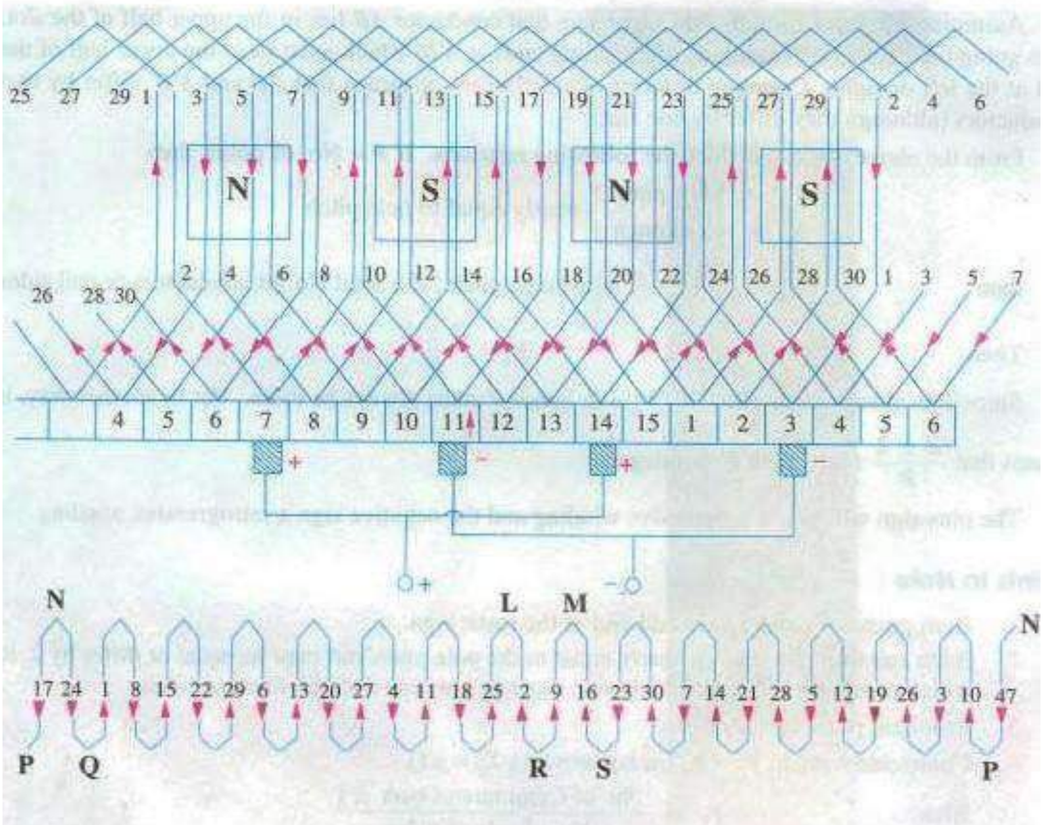


Fig. 26.36

As shown in fig. 26.36 and 26.37, conductor No. 5 is taken to conductor No. 5 4 7 12 at the back and is joined to commutator segment 5 at the front. Next. the conductor No. 12 is joined to commutatorSegment5 + 7 = 12 (Yc = 7) to which isjoined conductor No. 12+ 7 = 19. Continuing this Way, we come back NO. 5from where We started. Hence. the winding Closes itself.

If we take S, then the pitches would be : Yn 9 and YF 7 or Yb = 7 and YF 9. Incidentally, if YA = Yc is taken as 7, art-nature Vill rotate in one direction and if Yc = 8, it Will rotate in the Opposite direction.

The simple winding table is as under ;

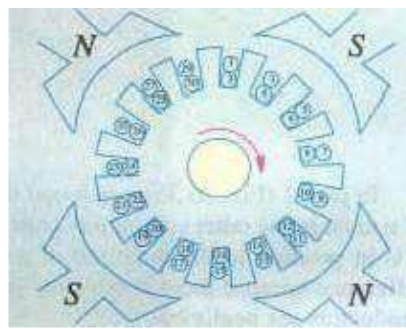
Back Connections Front Connections

1 to (1 + 7) = 8	$\longrightarrow$	8 to (8 + 7) = 15
15 to (15 + 7) = 22	$\longrightarrow$	22 to (22 + 7) = 29
29 to (29 + 7) = 36 = (36 − 30) = 6	$\longrightarrow$	6 to (6 + 7) = 13
13 to (13 + 7) = 20	$\longrightarrow$	20 to (20 + 7) = 27
27 to (27 + 7) = 34 = (34 − 30) = 4	$\longrightarrow$	4 to (4 + 7) = 11
11 to (11 + 7) = 18	$\longrightarrow$	18 to (18 + 7) = 25
25 to (25 + 7) = 32 = (32 − 30) = 2	$\longrightarrow$	2 to (2 + 7) = 9
9 to (9 + 7) = 16	$\longrightarrow$	16 to (16 + 7) = 23
23 to (23 + 7) = 30	$\longrightarrow$	30 to (30 + 7) = 37 = (37 − 30) = 7
7 to (7 + 7) = 14	$\longrightarrow$	14 to (14 + 7) = 21
21 to (21 + 7) = 28	$\longrightarrow$	28 to (28 + 7) = 35 = (35 − 30) = 5
5 to (5 + 7) = 12	$\longrightarrow$	12 to (12 + 7) = 19
19 to (19 + 7) = 26	$\longrightarrow$	26 to (26 + 7) = 33 = (33 − 30) = 3
3 to (3 + 7) = 10	$\longrightarrow$	10 to (10 + 7) = 17
17 to (17 + 7) = 24	$\longrightarrow$	24 to (24 + 7) = 31 = (31 − 30) = 1

Sincewe come back to the conductor No. 1 from where we started, the winding gets closed at this

stage.

Brush Position



Location of brush position in wave-winding is slightly difficult. In Fig. 26.36 conductors are supposed to be moving from left to right over the polcs_ By applying Fleming's Right-hand rule. the directions of the induced e.m.fs in various armature conductors can be found. The directions shown in the figure have been found in this manner. In the lower part of Fig. 26.36 is shown the equivalent ring or spiral diagram which is very helpful in understanding the formation of various parallel paths in the armature. It is seen that the winding is electrically divided into two portions. One portion consists ofconductors lying between points N and L and the other of conductors lying

between N and M. In the first portion, the general trend of Fig. 26.37 the induced e.m.fs, is from left to right whereas in the second

portion it is from right to left. Hence, in general, there are only two parallel paths through the winding, so that two brushes are required, one positive and one negative.

From the equivalent ring diagram. it is seen that point N is the separatiqg point Of the e.m.fs. induced in the two portions of the winding. Hence, this fixes the position of the negative brush. But as it is at thc back and not at the commutator end of the armature, the negative brush has two alternative positions i. e, either at point P or Q. These points on the equivalent diagram correspond to commutator segments NO. 3 and ll.

Now, we will find the position of the positive brush. It is found that there are two meeting points of the induced e.m.fs. i.e. points L and M but both these points are at the back or non-commutator end of the armature. These two points are separated by one loop only. namely. the loop composed of conductors 2 and 9, hence the middle point R of this loop fixes the position of the positive brush. which should be placed in touch with commutator segment No. 7. We find that for one position Of the brush, there are two alternative positions for the —ve brush.

Taking the -eve brush at point R and negative brush at point p, the winding is seen to be divided into the following two paths.

In path 1 (Fig. 26.36) it is found that e.m.r. in conductor 9 is in opposition to the general trend ofe.m.fs. in the Other conductor"s comprising this path. Similarly, in path 2. the e.m.f. in conductor 2 is in position to the direction Of e.rn.fs. in the path as a whole. However, this will make no difference because these conductors lie almost in the interpolar gap and, therefore e.m.fs. in these conductors are negligible.

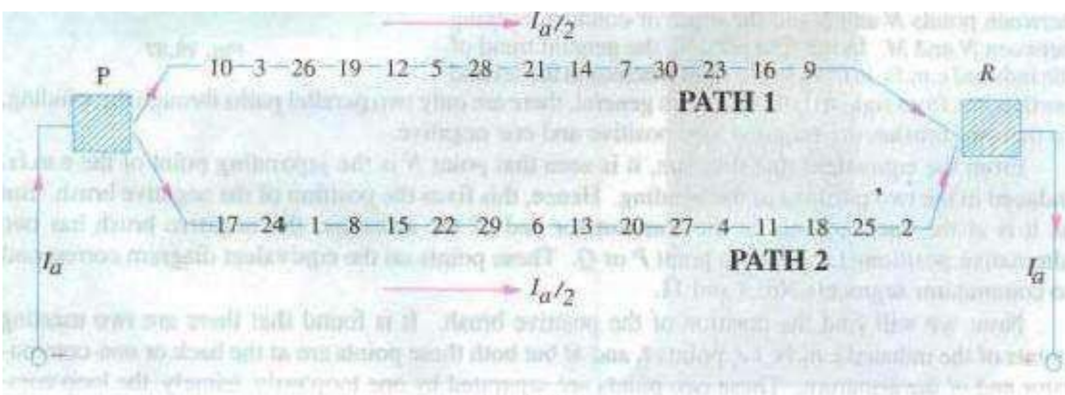


Fig.26Æ

Again, take the case of conductors 2and 9situated between points L and M. Since the armature conductors are in continuous motion over the pole faces, their positions as shown in the figure are only instantaneous. Keeping in this.nind. itis obvious that conductor 2 is about to move from the influence Of S-pole to that of the next N-pole. Hence, the e.m.f. in it is at the point Of reversing. However, conductor 9 has already passed the position of reversal, hence its e.m_f_ will not reverse. rather it will increase in magnitude gradually. It means that in a very short interval, point M will

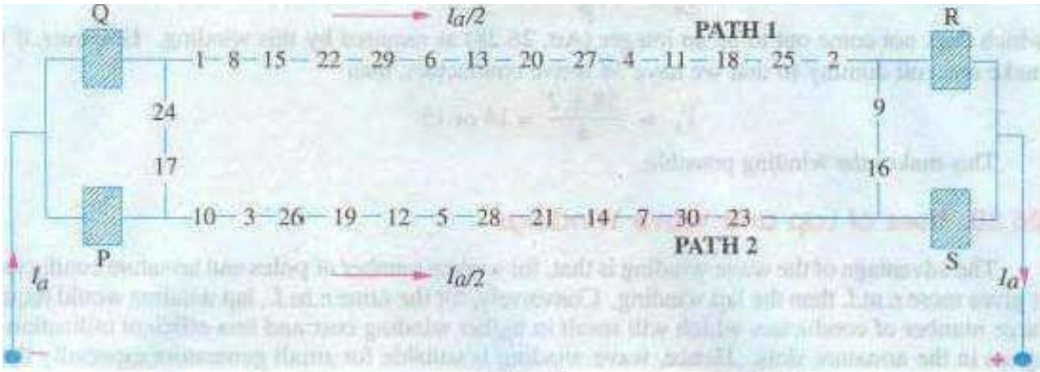


Fig. 26.40

become the meeting point of the c.m_fs. But as it lies at the back Of the armature, there are two alternative positions for the -Eve brush i.e. either point R which has already been considered or point Suhich corresponds to commutator segment 14. This is thesecond alternative position of the positive brush. Arguing in the same way. it can be shown that after another Short interval Of time. the alternative position Of the positive Will shift from segment 14 to segment 15. Therefore, if one positive brush is in the contact With segment 7, then the second positive bnlSh if used, should be in touch with both segments 14 and 15.

Il may be noted that if brushes are placed in both alternative positions for both positive and negative (i.e. if in 4 brushes are used, two -eve and two —ve), then the effect is merely to shortcircuit the loop lying between brushes of the same polarity. This is shown in Fig. 26.40 it Will also be noted that irrespective of whether only two or four brushes are used, the number of parallel pallLS through the armature winding is still two.

Summarizing the above facts, we get

1. Only two brushes are necessary. though their number may beequal tothe number of poles. 2. The number of paths through the armature winding is two irrespective Ofthe number of generator poles. That ig why this winding is sometimes called •two-circuit' or •series'

3. The generator e.m.f. is equal to the e.m.f. induced in any one ofthe two parallel paths. •If eav is the e.m.f. induced/conductor, then generator e.nl.f. is E x "2.

4. The equivalent armature resistance is nearly one-fourth Of the total resistance Of the armature winding,

5. If is the total armature current, then current carried byeach path orconductor is obviously 1/2 whatever the number of poles.

Dummy or Idle Coils

These are used with wave-winding and are resorted to when the requirements of the winding are not met by the standard armature punchings available in armature- winding shops. These dummy coils do not influence the electrical characteristics Of the winding because they are not connected to the commutator. They are exactly similar to the other coils except that their ends are cut short and taped. They are there simply to provide mechanical balance for the armature because an armature having some slots without windings would be out Of balance mechanically. For example. suppose number Of armature Slots is 15, each containing 4 sides and the number of poles is 4. For a simplex wave-windings, Dummy coils

60±2

4 which does not come out tobe an integer (Art. 26.28) as required by this winding. However, if we make one coil dummy so that we have 58 active conductors. then

58 ± 2

— 14 or 15 4

This makesthe winding possible.

Uses Of Lap and Wave Windings

The advantage of the wave winding is that, for a given number ofpoles and armature conductors, it gives more e.m.f. than the lap winding. Conversely, for the same e.m.f., lap winding would require large number Of conductors which will result in higher winding cost and less efficient utilization of space in the armature slots. Hence, wave winding is suitable for small generators especially those meant for 500-600 V circuits.

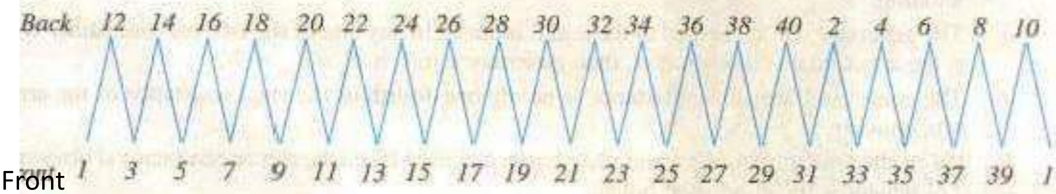
Another advantage is that in Wave winding. equalizing connections are not necessary whereas in a lap winding they definitely are. It is so because each Of the two paths contains conductors lying under all the poles whereas in lap-wound armatures, each of the P parallel paths contains conductors which lie under one pair of poles. Any inequality of pole fluxes affects two paths equally, hence their induced e.m.fs. are equal. In lap-wound armatures, unequal voltages are produced which set up a circulating current that produces sparking at brushes.

However, when laœe currents are required, it is necessary to use lap winding, because it gives more parallel paths.

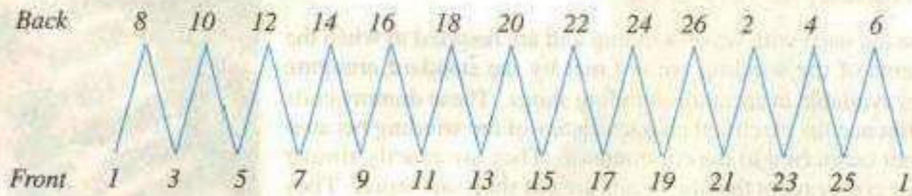
Hence, lap winding is suitable forcomparatively low-voltage but high-currentgenerators whereas wave-winding is used for high-voltage, low-current machines.

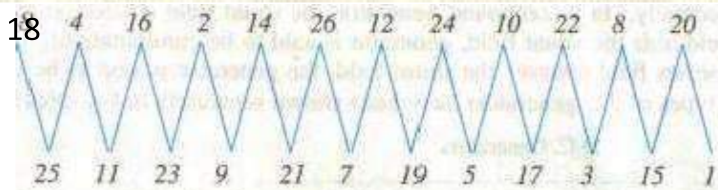
1. Write down the winding table for a 2-layer simplex lap-winding for a 4-pole d.c. generator having (a) 20 slots and (b) 13 slots. What are the back and front pitches as measured in terms of armature conductors ?

[Hint : (a) No. of conductors = 40 ; $Y_B = 11$ and $Y_F = -9$] (Elect. Engineering, Madras Univ. 1978)



(b) No. of conductors = 26 ; $Y_B = 7$; $Y_F = -5$





Tutorial Problem No. 26.1

simplex wave winding for a 4-pole d.c. machine with 28 conductors ? Explain

2. With a simplex 2-1ayer Wave winding having 26 conductors and 4-poles. write down the Winding table. What will be the front and back pitches of the winding ?

[Hint : $Y_F = 7$ and $Y_B = 5$]

(Electric Machinery-I, Madras Univ. Nov. 1979'

segments

Types of Generators

Generators are usually classified according to the way in which their fields are excited. Generators may be divided into separately-excited generators and self-excited generators.

(a) Separately-excited generators are those whose field magnets are energised from an independent external source Of d.c. current. It is shown diagramatically in fig. 26.41.

(b) Self-excited generators are those whose field magnets are energised by the current
(produce.d by the generators themselves. Due to residual magnetism, there is always present some flux in the When the armature is rotated, some and hence some induced current is produced which is partly or fully passed through the field coils thereby strengthening the residual pole flux,

There are three types of self-excited generators named according to the manner in which their field coils (or windings) are connected to the armature. lib Shunt wound

The field windings are connected across or in parallel with the armature conductors

Example 26.9. An 8•p01e dc. generator has 5(H) armature conductors, and a useful flux Of O. 05 Wb per pole. What Will be the e.m.f. generated ifit is lap-connected and runs at 4200 rpm ? What must be the speed at which is to be driven produce the same it is Wave-wound?

[U.P. Technical Univ. 2001)

916 Electrical Technology

Solution. With lap-winding, $P = a = 8$

$$E = \phi\,(N/60)\,(P/a)$$
$$= 0.05 \times 500 \times 20 \times 1,$$
$$= 500 \text{ volts}$$

for lap-winding

If it is wave-wound, $P = 8, a = 2, P/a = 4$

and $E = 0.05 \times 500 \times (N/60) \times 4$

For $E = 500$ volts, $N = 300$ rpm

Hence, with wave-winding, it must be driven at 300 rpm to generate 500 volts.

Additional Explanation. Assume 1 amp as the current per conductor.

(a) Lap-wound, 1200 rpm : 500 V per coil-group, 8 groups in parallel

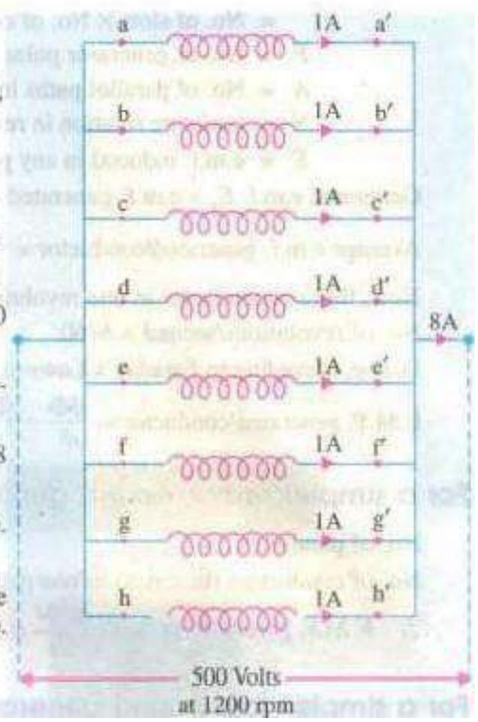
Net output current = 8 amp as in Fig. 26.51 *(a)*.

Power output = 4 kW

(b) Wave-wound, 300 rpm : 2 groups in parallel, one group has four coils in series, as shown in Fig. 26.51 *(b)*.

Total power-output is now

$$500 \times 2 = 1000 \text{ W}.$$



It is reduced to one fourth. being proportional to the

Example 26.26. A long-shunt dynamo, running ut 1000 voltage Of 220 V. The resistances

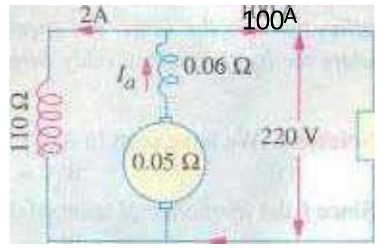
0.06 Q respectively. The overall efficiency at the above load is friction losses (c) the torque exerted by theprime,mover.

(Elect. Machinery-I. Bangalore Univ. 1987)

Solution. The generator is shown in Fig. 26.64.

220/110=2 A

1 -   100 A.



= 102 A

220 v

Drop in series field winding = 102 x   = 6.12 V

- 1022 xo.05 =520.2 W

Series field loss — 0.06 = 624.3 W Shunt field loss = 4k 110=440 w

Example 26.28. A long-shunt compound-wound generator gives 240 volts at EL output of 100 A. The resislances of various windings ofthe machine are : armature (including brush contact) O. J Ω series field 0.02 Q interpole field 0.025 shunt field (including regulating resistance' 100 Q. The iron loss EL is 1000 W ; windage andfriction losses 10101 SOO W. Calculate EL efficiency of the machine. i Electrical Machinery-I, Indot•e Univ. 1989)
output 240x

Total armature circuit resistance = 0.1 + 0.02 + 0.025 0.145 n

240/100=2ÅA 2.4= 102.4 A

Armature circuit   loss = 102.4^2 x O.   1,521 W

Shunt field copper loss = 2.4 x 240=576 W

Iron loss =   W ; Friction loss 500 W

24.000

Total loss = 1,521   -087= 87%

24.000 1597

DC MOTOR

Motor Principle

An Electric motor is a machine which converts electric energy into mechanical energy. ILS action is based on the principle that when a current-carrying conductor is placed in a magnetic field. it experiences a mechanical forec whose direction is given by Fleming's Left-hand Rule and whose magnitude is given by Constructionally, there is no basic difference between a d.c_ generator and a d_c_ motor In tact, the same d_c_ machine can bc used interchangeably as a generator or as a motor. D.C_ motors are also like generators. shunt-wound or series-wound or compound-wound.

of Motor are supplied with current from the supply mains. they expenencc a force tending to rotate the armature. Armature conductors under At-pole arc assumed carry current downwards (Crosses) and those under S-poles, to carry current upwards (dots). By applying Fleming's Left•harud Rule, the direction of the force on Fig.

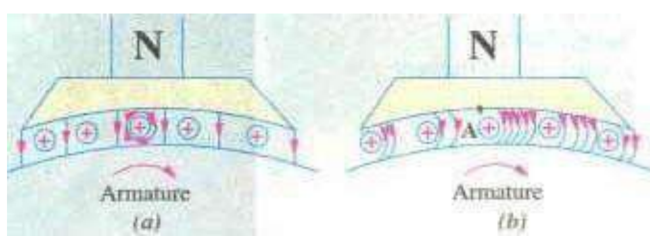
each conductor can found. It is shown by small arrows placed above each conductor. It Will be seen thåt each conductor can be found. It will be seen that each conductor experiences a force F which tends to mtate the armature in anticlcxkwise direction. These forces collectively produce a driving torque which sets the armature rotating.

It should be noted that the function of a commutatorin the motor is the same as in a generator. By revetsing current in each conductor as it passes from one pole toanother, ithelps to developa continuous and unidirectional torque.

Cornparison Of Generator and Motor

ActionAs said above, the same d-c.

machine



can be used. at least theoretical iy, interchangeably as a generator or as a motor. When operating as a generator, it is driven by a mechanical machine and it develops voltage which in turn produces a current flow in an electric circuit. When operating as a motor. it is supplied by electric Fig. 29.2 current and it develops torque which in turn produces mechanical rotation.

Lel us first consider its operation as a generator and see how exactly and through which agency, mechanical power is converted into electric power.

In Fig. 29.2 pan of a generator whose armature is being driven clockwise by its prime nlover is shown.

Fig, 29.2 (a) represents the fields set up independently by the main poles and the armature conductors like A in the figure. The resultant field or magnetic lines on flux are shown in Fig. 29_2 (b) It is Seen that there is a crowding of'iries of flux on the right-hand side ofA. These magnetic lines of flux may be likened to the rubber bands under tension. Hence. the bent lines Of flux up mechanical force on A much in the same way as the bent elastic rubber band Of a catapult produces a mechanical force on the stone piece. It Will he seen that this force is in a direction opposite to that Of armature rotation. Hence, it is known as backward force or magnetic drag on the conductors. It is against this drag action on all armat•uœconductor that the prime mover has to work. The WOEk done in overcoming this opposition is converted into electric energy. Therefore. it should be clearly understood thal it is only through the instrumentality of this magnetic drag that energy conversion is possible in a d.c.

generator.

Next, suppose thal the above d.c. machine is un• coupled from its prime mover and that current is sent through the armature conductols under a N.polc in the downward direction as shown in Fig. 29.3 The conduclors will again experience a forec in the anticlockwise direction ('Fleming's Left hand RAIle)_

Hence. the machine Will

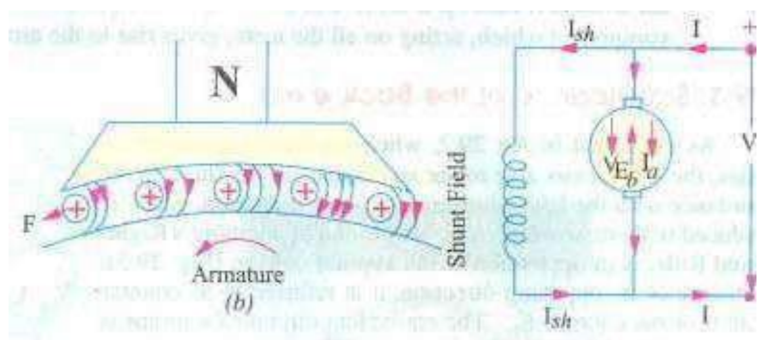
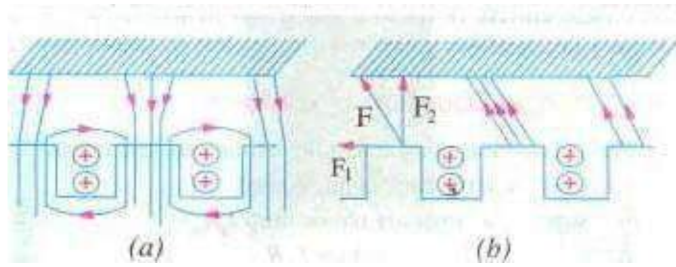


Fig. 29.3 (a) Fig. 29.3 (b)

Start rolating anticiockwise. thereby developingÆi torque which can produce mechanical rotation. The machine is then said to be motoring.

As said above, energy conversion is not possi'le unless there is some opposition whose over% coming provides the necessary means for such conversion. In the case Of a generator, it was the magnetic drag which provided the necessary opposition. But what is the equivalent of that drag in the case Of a motor ? Well, i' i' the hack e.rn,f. It is explained in this manner:

As soon as the armature starts rotating. dynamically (or motionally) induced e.m.f. is produced in the armatUre conductors.



The direction of this induced e.m.f. as found by Fleming's Right-hand Rule. is outwards i_eu. in dilect opposition to the applied voltage (Fig. 29.3 This is Why it is known as back e.m.f. or counter Its value is the Same as for the motionally induced e.mf in the generator i.e. 14, ('ÞZ'V'x (P/A) volts. The applied voltage Vhas to be forcecurrent through the armature- conductors against this back e_m.f. Eh. The electric work done in overcoming this opposition is convened into mechanical energy developed in the armature. Therefore. it is obvious that bulfor the pro. auction of this opposing e.m.f. energy conversion Would not have been possible.

Now. before leaving this topic, 29,4 Fig. let it be pointed out thai in an actual motor with slotted armature. the torque is not due to mechanical force on the conductors themselves, but due to tangential pull on the armature teeth as shown in Fig. 29.4.

It is seen fig. 29.4 that the main flux is concentrated in the form of tufts at the armature teeth while the armature is shown by the dotted lines embracing the armature slots. The effect of

• fact, it seems to he one of the fundamental laws or no energy conve4Sion from one to another is ['ossible until there is some one to oppose the conversion. Hut forthe presence of this opposition, there would simply t-,c no energy conversion. Jn generators. opposition is provided by magnetic drag whereas in motors. back c.m.f. does this job. Moreover, it is only that part of the input energy which is used for overcoming this Opposition that is converted into the Other form.

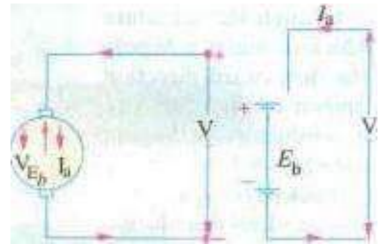
-armature flux on the main flux, as shown in Fig. 29.4 is two-fold ,

(i) It increases the flux on the left-hand side of the teeth and decreases it on the right-hand side, thus making the distribution Of flux density across the tooth section unequal.

(ii) it inclines the direction Of lines Of force in the air-gap so that they are not radial but are disposed in a manner shown in Fig. 29.4 The pull exerted by the poles on the teeth can now he resolved into two components. One is the tangential component Fl and the other vertical component e. The vertical component Fl. when considered for all the teeth round the armature, up zero. But the comlN)nent FI is not cancelled and it is this tangential component which, acting on all the teeth, gives rise to the armature torque.

Significance of the Back e.m.f.

As explained in Art 29.2, When the motor armature rotales, the conductors also rotate and hence cut the flux. In ac. cordanee with the laws of electromagnetic induction, emi,f. is induced in them whose direction, as found by Fleming's Righthand Rule. is in opposition to the applied voltage (Fig. 29.5). Becáusc Of its opposing direction. it is referred to as counter• e,rn.f. or back



e.m.f. Eh. The equivalent circuit of a motor is shown in Fig. 29.6. The rotating armature generating the hack e. m.f. Eh is like battery Of e.m.f. Eh put across a supply mains of V volts. Obviously, V has to drive la against the opposition Fig, 29.5 of Eh. The power required to overcome thig opposition is Eh/n.

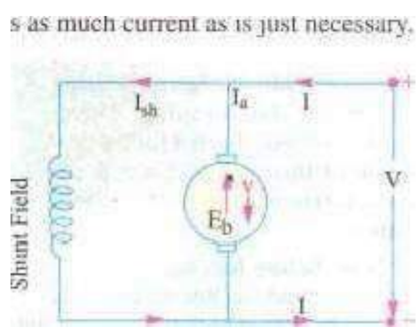
In the case of a cell. (his power over an inter,'al of time is converted into chemical energy, but in the present case, it is converted into mechanical energy.

It will be seen that I Net volta½ _ $V - V_b$

Resistance Ra

where R is the resistance Of the armature circuit. As pointed out above, Eh = CÞZV (P/A) volt where N is in r.p.s.

Back e_m.f_ depends, among other faclors. upon the armatuœ speed. If speed is high. Eb is large, hence armature cunent seen from the above equation, is small. If the speed is less. then Eb is less, hence more current flows which develops motor (Art 29.7). so. we find that Eh acts like a governor i.e„ it makes a motor self-regulating so that it draws as much as is just necessary.



Voltage Equation Of a Motor

The voltage V applied across the motor armature has to overcome the back e.m.f. Ebund in supply the armature ohmic drop

This is known as voltage equation of a motor. Now, multiplying both sides by , We get

As shown in Fig. 29.6, Fig.2g.6

V = Eectrical input to the armature

Eh/o = Electrical equivalent Of mechanical power developed in the armature I: R , = Cu loss in the armature

Hence. out Of the arunature input, some is wasted in 1^2R loss and the rest is converted into me— chanical power within the armature.

ltmay also be noted that motor efficiency is given by the ratio of power developed by the arma-

ture to its input i.e. EBIJV la = E*/V. Obviously. higher the value of Eb compared to V. higher the motor efficiency.

Condition for Maximum Power

The gross mechanical power developed hy a motor is pm = la — I: Ra.

Differentiating both sides with to la and equating the result to zero. we get v—2/aRa-O As and

Thus gross mechanical power developed by a motor is maximum When back e.tn_f. is equal to half the applied voltage, This condition is, however. not realized in practice, because in that ease current would be much beyond the normal current of the motor. Moreover. half the input would be wasted in the form of heat and taking other losses (mechanical and magnetic) into consideration. the motor efficiency will be well below 50 percent-

Example 29.1. A 220-V dc. machine has an armature resistance Of OS the full-load

armature is 20A,find acts as motor

1987)

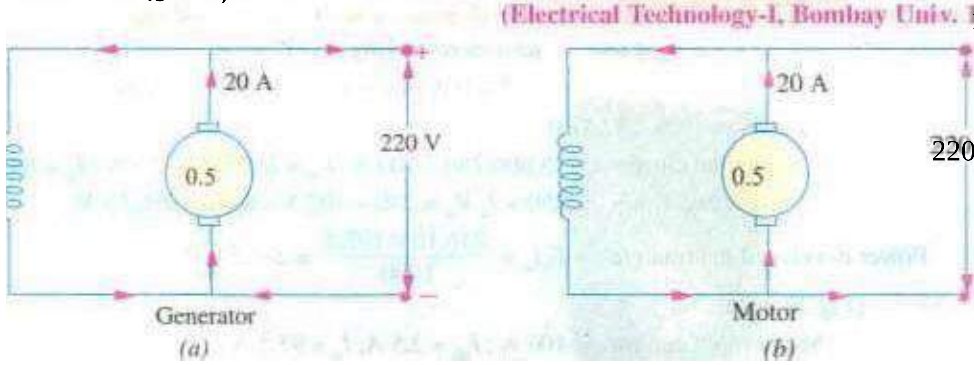


Fig.29.7

Solution. As shown in Fig. 29.7, the machine is assumed to be shunt•connected. In each ease, shunt current is considered negligible because its value is not given.

As Generator [Fig. V

Example 29.2. A separately excited D, C. generaror has armature circuit resistance OfO. I Ohm and the brush-drop is 2 V. When running at 1000 r.p.m., it delivers a of 100 A at 250 V to a load ofconstant resistance If the generator speed drop to 700 cp. m., unaltered, find the current delivered load. AMIE, Electrical Machines, 2001) solution.

At = 262 X 700/1000 = 183.4 V If is the new current. — 2 — = 2.5

•This gives 96.77 amp.

Extension to Question : With what loudresistance will the current be amp. at 700 r.p.m. ?

Solution.

Forlo- IOOamp.and E 183.4 v, RL- 1.714 ohms.

Example 29.3. A shunt motor has armature resistance Of 0.8 n andfield resistance Of

200 Determine the back e.m.f. When giving an output of7.46 kW at 85 pert-ent efficiency,

Solution. Motor input 7.46 x 10^3/0.85 W

Motor input current $= 7460/0.85 \times 440 = 19.95$ A ; $I_{sh} = 440/200 = 2.2$ A

$I_a = 19.95 - 2.2 = 17.75$ A ; Now, $E_b = V - I_a R_a$

$\therefore \quad E_b = 440 - (17.75 \times 0.8) = 425.8$ V

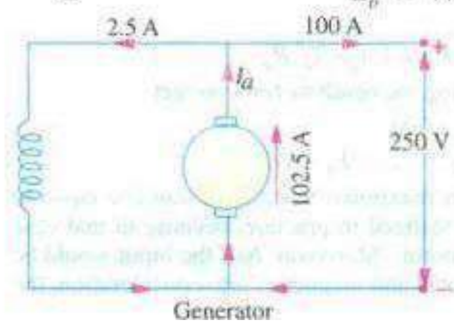


Fig. 29.8 (a)

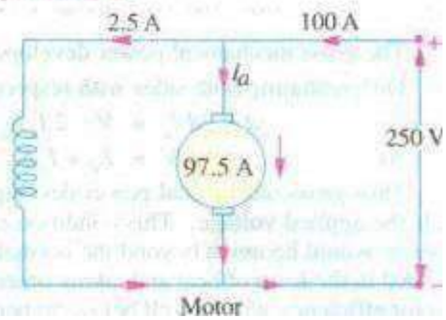


Fig. 29.8(b)

Example 29.4. A 25-kW. dc. shunt generator has armature andfield resistances of 0.06 Q and 100 respectively Determine the total armature power dewloped when working (i) as a generator delivering 25 kW output and t ii) as a motor faking 25 kW input.

' Electrical •lèchnoloo, Punjab Univ„ June 1991

Solution. As Generator (Fig. 29.8 (a)/

output current = 25 ,000/250 =100 A $I_{sh} = 250/100 = 2.5$ A ; $I_a =$ 102.5 A

Generated emm = 250+1 R„-250+ 102.5 xo.06=256.15 V

256.15 x 102.5

Power developed in armature

As Motor (Fig

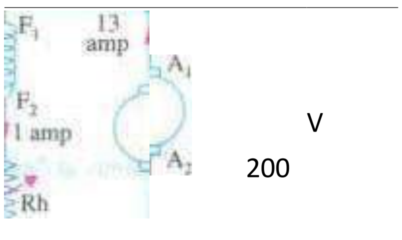
Motor input current

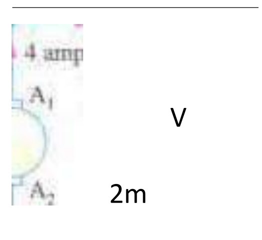
244.15 v

Power in armature = kW

Example shunt generator With terminal voltage of 200 volts delivering 12 amps the load 200 ohms. is driven at 1000 Calculate the flux per pole in the machine. If the machine has to be run as motor With the same terminal voltage and drawing 5 amps from the mains, maintaining the same magneticfield, find the speed of the machine, [Sambalpur University. 19981

Soluti"n. Current distributions during two actions are indicated in Fig. 29.9 (a) and As a





generator, 1 3 amp

12 amp 5 amp

Supply

(a) Generator-action (b) Motor-action

Fig. 2"

$E_g = 200 + 13 \times$ 2=226 V

$\phi \frac{ZN}{60} \times \frac{P}{a} = 226$ 60 a

For a Lap-wound $P = a$ armature,

0.42375 wb $\phi = \frac{226 \times 60}{1000 \times 32} =$

Asa motor, $I_a = 4$ amp

$E_b =$ 200-4x2—192V

Giving N $= \phi ZN/60$

0.42375 32 $N = \frac{60 \times 192}{0.42375 \times 32}$

=850 r.p,m.

Tutorial Problems 29.1

I. What do you understand by the term 'back c,rn.f.• Ad.e_ supply has an armature resistance of O. IS Q, Calculate

(a) The Of hack e,m.f. when the armature current is 120 A.

The value of armature current when the hack is 4414 V.

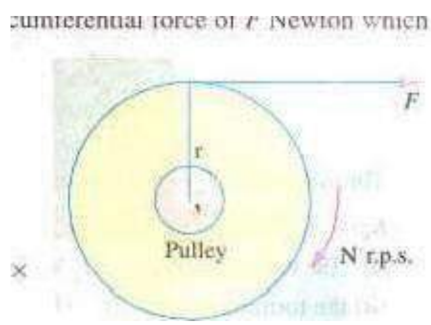
2. A d.c. motor connected to 460-V supply utkes an armaturc Current Of 120 A on full load. If the circuit has a of 0.25 Q, calculate the value of the back c.m.f. at this load. | •130 VI A 4•poIe d:c_ motor takes an armature current Of 150 A at

3. resislance 440 V. If its armature circuit has a of O. IS Q. what will the value Of back c.rn.f, at this load ? 1417.5 VI

Torque

By the term torque is meant the turning or twisting moment of a force about an axis. It is measured by the product of the force and the radius at which this force acts.

Consider a pulley Of radius r metre acted upon by acircumfcrential force of F Newton which Causes i' to rotate at N r.p.rn. (Fig. 29.10).



Then torque = Ex r Newton-metre(N - m)

Work done by this force in one revolution

= Force distance = F x 2rtr Joule Power developed = F x 2 rtrx N Joulusecond or Watt

Now 2 lt.'V Angular velocity in radian/second and F

—Torque r

∴Power developed = Tx (D watt or P T Watt if N isin then Fig. 29.10

Moreover, m = 2 wV/60 rad's

2 It,'V

= $\times T$ or P— 60

Armature Torque Of a Motor

Let Ta bethe torque developed by the armature of a motor running at N r.p.s• If is in MM, then power developed Tu x N watt

We also know that electrical Ix»ver converted into mechanical Il'w•er in the armature Alt 29.4) — Eh/a Watt

In the case of a series motor,

Windings carry full armature Current

(b) For shunt motors, Φ is practically constant, hence $T_a \propto I_a$

As seen from (iii) above

$$T_a = \frac{E_b I_a}{2\pi N} \text{ N - m - N in r.p.s.}$$

If N is in r.p.m., then

$$T_a = \frac{E_b I_a}{2\pi N/60} = 60 \frac{E_b I_a}{2\pi N} = \frac{60}{2\pi} \frac{E_b I_a}{N} = 9.55 \frac{E_b I_a}{N} \text{ N-m}$$

Shaft Torque (

The whole of the armature torque, as calculated is not available for doing useful work. because a certain percentage of it is required for supplying iron and friction losses in the motor.

The torque Which is available for doing useful work is known as Shaft torque It is so called because it is available at the shaft_ The motor output is given by Output x2mV Watt provided Tsh is in N-m and N in r.p,s.

Output in watts $T_{sh} = \frac{\text{Output in watts}}{2\pi N}$ N- m −N in r.p.S

$$= \frac{\text{Output in watts}}{2\pi N/60} \text{ N-m} - N \text{ in r.p.m.}$$

$$= \frac{60}{2\pi} \frac{\text{output}}{N} = 9.55 \frac{\text{Output}}{N} \text{ N-m.}$$

The difference (Ta — T,h) is known as lost torque and is due to iron and friction tosses of the motor.

Note. The value of back e.m.f. Eh can be found from the equation, Eb = V — la Ra

(iii the formula Eb = Z.'V x (P/A) volt

Example 29.6. A dc. motor takes an armature current Of at 480 V! The armature circuit resistance is 0.2 Q The machine has 6•poies and the armature is lap-connected with 864 conductors. The flu per pole is 0.05 Wh. Calculate the speed and the gross torque developed by the

(i)

armature. (Elect. Machines, A.M.I.E. Sec B. 1989)

Solution. Eh 480—110 xo.2—458V,

0.05 x 864 x N 6

60

N 636 cp.m.

T. = 0.159 xo.05

Example 29.7. A 25CL •a-pole, wave-wound Series motor has 782 conductors on its armature. If has armature and seriesfield resisiance Of0.75 Ohm. The motor takes a currvnt A. Estimate its speed and gmss to rque developed if it has a per pole of 25 m Wb.

(Elect. Engg.•ll, pone Univ. 1991)

Solution.

Example 29.8. A dc. Shunt Find torque mechanical power developed for an armature current Of 50 A. State the simplifying assumptions. (Basic Elect. Machine Nagpur Univ. its 1993) Solution. A given d_c. machine develops the Same e.m.f. in its armature conductors whether running as a generator or us a motor, Only difference is that this armaturee.m.f. is known as back

e.m.f. When the machine is running as a motor.

Mechanical power developed in the arm 50 = 12.500 W

T. —9.55 9.55 x 250 79.0 N-m.

Example 29.9. Determine developed and shaft torque of220-V,4-pole series motor with 800 conductors Wove-connected supplying load 0/82 kW by taking45 Afrom the mains. Theflltv per pole is 25 m Wh and its armature circuit resistance is 0.6 Q.

(Elect. Machine AMIE sec. B Winter 1991)

. N2/500=200/210

Example 29.11. -500-V.37.3 kW. 1000 r.p.m. d,c. shunt motor has on full-load an efficiency OJ 90 The annature circuit resistance 0.24 Q and there is Iola/ voltage drop Of 2 V al the brushes. The field curren' is 1.8 A. Determine (i) full-load line curren/ (in full load shaft torque in N-m and resistance in motor starter 'o limit the starting cumm to 1.5 rimes the full-load current. (Elect. Engg. l; M.S. Univ. Baroda 1987)

Solution. Motor input W = 37,300/0.9 = 41,444

= 41,444/500 = 82.9 A

F.l. line current = 41.444/S(.WJ

= 356N-m

If R is the Starter resistance (which is in series with armature), then

Example 29.12. A "-pole. 220- V shunt motor has 540 lap-wound conductor. rakes 32 A from the supply nminsanddevelops output power of5.595 kW Thefield winding takes A. The resistance is 0.09 Q and the per pole is 30 mWb_ Calculate the speed and rhe torque developed in newton-metre. (Electrical Nagpur Univ. 1992'

Solution. = 31 A V

,ÞZN P

Now. 217.2 =

60 A 60 N = 804.4 r.p.m.

T. 9.55 x output in watts = 9.55 x 595 = 66.5 N-m

804.4

Example 29.13 Find the lead andfull-load speedsfor afour-pole, 220-V. and 20-kW. shunt motor having thefollowing data :

Field—current = 5 amp, armature resistance = Ohm.

Flux per pole 0.04 Wb. number ofarmature-conductors = 160, livo-citèuit wave-connection, full load current = 95 amp. No load current 29 A. Neglect armature reaction.

(Bharathitbasan Univ. April 1997)

Solution. The machine draws a supply current Of 9 amp at no load. Out Of this, 5 amps are required for the field hence the armature carries a no-load current of amp.

At load, armature-current is 90 amp. The armature-resistance-drop increases and the back e,m.f. decreases. resulting into decrease in speed under load compared to that at No-Load. : E., = — 4 x 0.04 = volts

Substituting this.

004 x 160 x

No-Load speed. No — 1030.5 r.p.m.

Armature current A, Ea — 90 x 0.04 216.4 V (216.4/219.84) x 1030.5 = 1014.4 rpm.

F•ample 29. i 3 Armature Ofa 6-pole. 6-circuit D.C. shunt motor takes A at a speed Of 350 r.p.m_ Theflux per pole is 80 milli-webers. the number Ofarmature rums is 600. and ofthe torque is lost in windage. friction and iron-loss. Calculate the brake-horse-power.

(Manonmaniam Sundaranar Univ. NOV. 1998)

Solution. Number of armature turns = 600

Therefore, Z = Number Of armature conductors = 1200 If electromagnetic torque developed is TNw — m.

Armature 'X'Wer T T x 2 't

= Twatts

To calculate armature power in terms of Electrical parameters, E must be known.

- SOX x

— 560 volts

With the armature current Of 400 Armature power = 560* 400 watts Equating the two,

T = 560 x 100/36.67—6108-5 Nw—m, Since 3 % of this torque is required for overcoming different loss-terms,

n-

Net torque = 0.97 x 6180.5 = 5925 Nw-m

For Brake-Horse-power, net output in kW should be computed first. Then ••kW" is to be co vetted to "BHV'. with 1 HP = 0.746 kw,

Net output in kW = 5925 36.67 10-3 —217.27 kW

Converting this to BHP- the Output = 291 _25 HP

Example 29.13 Determine the torque established by the armature Of a four-pole D.C motor having 774 conductors, two pa/hs in parallel, 24 milli-webers ofpole-flux and the armature current is 50 Amps. (Bharathiar Univ. April 1998)

Solution. Expression for torque in terms Of the parameters concerned in this problem is as follows :

T = 0.1590ZL„p/aNw-m Two paths in parallel for a 4-pole case means a wave winding.

T 0.159 X (24 x 774x50x4/2

= 29536 Nvv-m

Example 29.13 A 500-V D.C shunt motor draws a line-current of 5 A on tight-load. If armature resistance is O. IS ohm and field resistance is 200 ohms, determine the efficiency of the machine running as a generator delivering load current Of40 Amps.

gBharathiar Univ. April 1998)

Solution. (i) No Load. running as a motor :

Input Power = x 5 = 2500 watts field copper-loss = 500 x 2.5= 1250 watts

Neglecting armature copper-loss at no load (since it comes out to be 2.5^2 x 0.15 I watt). the balance of 1250 watts of power goes towards no load losses of the machine running at rated speed, These losses are mainly the no load mechanical losses and the core-loss.

(ii) As a Generator, delivering 40 A 10 load :

Output delivered 500x40x10^{-3} = 20 kW

Losses : Field copper-loss = 1250 watts

(b' Armature copper-loss = 42.5' x = 271 watts

(c) NO load losses = 1250 watts

Total losses kW

Generator Efficiency (20/22.771) x 100 % = 87.83 %

Extension to the Question : At Bihar speed should the Generator bé run, ifthe shunt-field is not changed, in the above case ? Assume that the motor was running 600 r,p.m, Neglect armature'

Solution. As a motor on no-load.

$E_{b0} = 500 - I_a r_a = 500 - 0.15 \times 2.5 = 499.625$ V

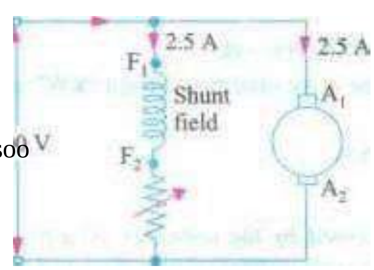
As a Generator With an armature culTent Of 42.5 A.

E_{b0} =— 500+42.5xo.15 -506.375V

Since, the terminal voltage is same in both the cases. shunt field current remains as 2.5 amp. With armature reaction is ignored, the flux/pole remains same. The e.m_f. then becomes proportional to the speed. lithe generator must driven at N rp.m.

N = (506-375/449.625) 608.1 r.p.m.

23 A 40 A



(a) Motor at no load (b) Generator loaded

Fig. 29.11

Note. Alternative to this slight increase in the speed is to increase field current with the help of decreasing the rheostatic resistance in the field-circuit.

Example 29.13 A dc. series motor takes 40 A at 220 V and runs at 800 r.p.m- If the armature andfield res/stancc are 0.2 Q and O. I Q respectively and the iron andfriction are O.S W find the torque developed in the armature. What Will be the output Ofthe motor ?

Solution. Armature torque is given by $T_a = 9.55 \dfrac{E_b I_a}{N}$ N-m is given by Ta = 9.55

$E_b = V - I_a (R_a + R_{se}) = 220 - 40\,(0.2 + 0.1) = 208$ V

$T_a = 9.55 \times 208 \times 40/800 = 99.3$ N-m

Cu toss in armature and ies-field resistance $= 40^2 \times 0.3 = 480$ W series-field resistance = 40^2 0.3

[Ton and friction losses = SOO W Total losses 480 + 500 = 980 W

Motor power input = 220 40 = 8.800 W

Motor output 8.800 - 7.820 W = 7.82 kW

Example 29.14. A curring tool exerts a tangential force of 400 N on a steel bar of diameter Which is being t"rned in a simple lathe. The lathe is driven by a chain at 840 rpm. from a 220 V dc. Motor which runs at 1800 r.p.m- Culc•ulule the current taken by the motor ifits efficiency is 80 Whal size is the motor pulley ifthe lathe pulley has a diameter of24 cm .

(Elect. Technology-Il, G"aiior Univ. 1985)

Solution. Torque Tangential force x radius = 400 x 0.05 = 20•N-m

Output power = T. x 27th' watt 20 x x (840/60) watt = I W

Motor Motor input - 1,760/0.8 - 2,200 W

Current drawn by motor = 2200/220 10 A

Let N, and be the speed and diameter of the driver pulley respectively and N2 and t): the respective speed and diameter of the lathe pulley.

10 cm

N, x D, or

Example 29.1 S. The armature winding Ofa 200-V.4-pole. series motor is lap-connected. There Ire 280 slots and each slog has conductors. The current is 45 and thejit" per pole is 18 m W/O.

The field resistance is 0.3 Q; resistance and the iron and friction losses total W. The pulley diameter is O. m. Find the pull in newton at the rim of the pulley.

(Elect. Engg. Sec. A. 1991'

900

Solution. Total input $E_b = V - I_a R_a = 200 - 45(0.5 + 0.3) = 164$ V

Now Iron + $E_b = \frac{\Phi ZN}{60} \cdot \left(\frac{P}{A}\right)$ volt

Friction losses

Output $164 = \frac{18 \times 10^{-3} \times 280 \times 4 \times N}{60} \times \frac{4}{4} \quad \therefore \quad N = 488$ r.p.m.

= 9 x 55 x 6580 - 12SN-m

488

...ut = $200 \times 45 = 9{,}000$ W ; Cu loss = $I_a^2 R_a = 45^2 \times 0.8 = 1{,}620$ W

...ses = 800 W ; Total losses = 1,620 + 800 = 2,420 W

...ut = 9,000 − 2,420 = 6,580 W

Let F be the pull in newtons at the rim of the pulley.

FxO.205 = 128.8 F = 128.8/0.205 N = 634 N

Example 29.16. A 4-pole, 240 V. wave connected shunt motor gives 1119 kW when running at 1000 r.p.m. and drawing armature and field currents of 50 A and A respectively. has 540 conductors. *Its resistance is 0.1 Ω. Assuming a drop of 1 volt per brush, find (a) total torque (b)* useful *torque (c) useful flux / pole (d) rotational losses and (e) efficiency.*

Solution. $E_b = V - I_a R_a -$ brush drop $= 240 - (50 \times 0.1) - 2 = 233$ V

Also $I_a = 50$ A

(a) Armature torque $T_a = 9.55 \frac{E_b I_a}{N}$ N-m $= 9.55 \times \frac{233 \times 50}{1000}$ = **111 N-m**

(b) $T_{sh} = 9.55 \frac{\text{output}}{N} = 9.55 \times \frac{11{,}190}{1000}$ = **106.9 N-m**

(c) $E_b = \frac{\Phi ZN}{60} \times \left(\frac{P}{A}\right)$ volt

$\therefore \quad 233 = \frac{\Phi \times 540 \times 1000}{60} \times \left(\frac{4}{2}\right) \quad \therefore \quad$ **$\Phi = 12.9$ mWb**

(d) Armature input = $V I_a = 240 \times 50 =$ **12,000 W**

Armature Cu loss = $I_a^2 R_a = 50^2 \times 0.1 = 250$ W ; Brush contact loss = $50 \times 2 = 100$ W

$\therefore$ Power developed = 12,000 − 350 = 11,650 W ; Output = 11.19 kW = 11,190 W

$\therefore$ Rotational losses = 11,650 − 11,190 = **460 W**

(e) Total motor input = $VI = 240 \times 51 = 12{,}340$ W ; Motor output = 11,190 W

$\therefore$ Efficiency = $\frac{11{,}190}{12240} \times 100$ = **91.4 %**

Example 29.17. A 460-Vseries motor runs at 500 rpm. taking a currergof40A. Calculate the speed and percentage change in torque if load is reduced so the is taking 30 A. Total resistance Ofthe armature andfield circuits is 0.8 Assume,i71Lx is pmportional to thefield current.

(Elect. Engg.-lt. Kerala Univ. 1988)

Solution, Since $\Phi \propto I_a$, hence $T \propto I_a^2$

$\therefore \quad T_1 \propto 40^2$ and $T_2 \propto 30^2 \quad \therefore \quad \frac{T_2}{T_1} = \frac{9}{16}$ T, Example 29. IS. A 460-1'.'55.95 750 r.p.m. shunt motor drives a load having a moment Of

2

inertia Of 252.8 kg-m. Find approximate time to attain full speed When starting from against full-load torque if starting current varies between 1.4 and /.8 timesfull-load current.

Solution. Let us suppose that the starting current has a steady value of + 1.8)/2 = 1.6 times full-load value.

Full-load output 55.95 kW = 55,950 W Speed 750 = 12—5 r.p-S.

Fl.. shaft torque T = power/co = power/21tN = 55,950 x (750/60) 712.4 N-m

During starting period, average available torque

— 1.6

This torque acts on the moment of inertial I 252.8 km-rn:.

12.5

427,4 = 252.8 x 252.8 x dl 46.4 s

Example 29.19. A 14.92 kW 400 V. 400 -rpm. d.c. shunt motor draws a current of40A when running atfull-load. The moment ofinertia ofthe rotating system is 7.5 kg-m^2. Ifthe starring current is .2 rimesfull-load current. calculate (a) full•load torque fbj the time requiredfor the motor ro attain the rated speed against full-load.

Gujarat Univ. 1988)

Solution. FL output 14.92 kW = 14,920 W ;

Now. Tm-output 14.920/2-„x

During the starting period, the torque available for accelerating the motor armature is

T—O.2x356-71.2N.m

NOW, torque I dt 4.41 second

Speed Of a D.C. Motor

From the voltage equation of a motor (An. 27.4), we get

$$E_b = V - I_a R_a \quad \text{or} \quad \frac{\Phi Z N}{60}\left(\frac{P}{A}\right) = V - I_a R_a$$

$$N = \frac{V - I_a R_a}{\Phi} \times \left(\frac{60A}{ZP}\right) \text{ r.p.m.}$$

$$V - I_a R_a = E_b \quad \therefore \quad N = \frac{E_b}{\Phi} \times \left(\frac{60A}{ZP}\right) \text{ r.p.m. or } N = K\,\frac{E_b}{\Phi}$$

It shows that speed is directly proportional to back can-f. Eb and inversely 10 the flux Φ

Now

$N \propto E_b/\Phi$. on

For Series Motor

Let N_1 = Speedin the 1st case : la, armature current in the 1st case = Φ_1 tluxlpole in the first cwse

I_{a2}, Φ_2 = corresponding quantities in the 2nd case.

Then. using the above relation, we get

$$N_1 \propto \frac{E_{b1}}{\Phi_1} \text{ where } E_{b1} = V - I_{a1} R_a \,;\, N_2 \propto \frac{E_{b2}}{\Phi_2} \text{ where } E_{b2} = V - I_{a2} R_a$$

$$\frac{N_2}{N_1} = \frac{E_{b2}}{E_{b1}} \times \frac{\Phi_1}{\Phi_2}$$

prior to saturation of magnetic poles ; $\Phi \propto I_a \quad \therefore \quad \frac{N_2}{N_1} = \frac{E_{b2}}{E_{b1}} \times \frac{I_{a1}}{I_{a2}}$

For Shunt Motor

In this case the same equation applies.

$$\frac{N_2}{N_1} = \frac{E_{b2}}{E_{b1}} \times \frac{\Phi_1}{\Phi_2} \qquad \text{If } \Phi_2 = \Phi_1, \text{ then } \frac{N_2}{N_1} = \frac{E_{b2}}{E_{b1}}.$$

Speed Regulation

The term speed regulation refers to the change in speed Of a motor With change in applied load torque. other conditions remaining constant. By change in speed here is meant the change which occurs under these conditions due to inherent properties of the motor itself and not thosc changes which are affected through manipulation of rheostats or other speed-controlling devices.

The speed regulation is defined ws the change in speed when the load is reduced from ruled value ro zero, expressed a.s percent Of the rated load speed.

% speed regulation = NL speed — FL speed x

F.L srwed

Torque and Speed of a D.C. Motor

It will be proved that though torque of a motor is admittedly a function of flux and armature Current. yet it is independent speed. In fact. it speed Which on torque and not vice-versa. It has proved earlier that

...Art. 27.9

..Art 27.7

It is seen from above that increase in would decrease the speed but increase the armature torque. It cannot be so because torque aluays tends to produce rotation. If torque increases, motor speed must increu•e rather than decrease. The apparent inconsistency between the above two equations can be reconciled in the following Way :

Suppose that the flux of a motor is decreased by decreasing the field current, Then, following sequence of events take place:

l. Back e.m_f_ Eb ,VCÞ/K) drops instantly (the speed remains constant because of inertia of the heavy armature).

2. Due to decrease in Eh, is increased because l. = (V— Moreover. asmall reduction in flux produces a proportionately large increase in armature current.

3. Hence, the equation Ta a small decrease in is more than counterbalanced by a large increase in with the result that there is a net increase in To.

4• This increase in ra produces an increase in motor speed.

It is seen from above that with the applied voltage V held constant, motor speed varies inversely as the flux. However, it is possible to increase and, at time, increxse the speed provided I_a is held constant as is actually done in a d.c. servomotor.

Example 29.20. A 4-pole series motor has 944 wave-connected armature conductors. At a certain load, the per pole is 34.6 and the total mechanical torque developed is 209 N-m. Calculate the line current laken by the motor and the speed at which it Will run With an applied voltage Of500 V Total motor resistance is 3 Ohm

(Elect. Engg. See A Part 11 June 1991

Solution. 0.1590 Z/a (P/A) N-m

209 — 0.159 x 34.6* 10-3 *944 x 1, (4/2); la = 20.1 A

as given in Art.

10-3x944xNx2

Example 29.21. A 25th V' shunt motor runs at 1000 r.p.m. at no-load and takes 84. The total armature and shuntfield resistances are respectively 0.2£2 and 250 Q. Calculate the speed when loaded and raking 50 A. Assume the

—240.2 V

2402

1000 248.6

Example 29.22. A dc. series motor operates at 800 r.p.m. with a line current of 100 A from

Example 29.23. A 230- Vd.c. shunt motorhas an armature resistance OfO_5 Q andfield resistance Of 115 Q At no load. the speed is 1.200 r.p,m. and the armature current 2.5 A On application Of rated load. the speed drops to J, 120 r.p.m. Determine rhe line current and power input when rhe motor delivers rated load. (Elect. 'Iëchnoloxy, Kerala Univ. 1988b 1986)

Solulion.

Line current drawn by motor Power input at rated load

Esample 29.24. A belt-driven shunt generator running 300 r.p.m. on 220- V bus-bars continues to run as a motor When the belf breaks, then taking k W. What Will be its speed ? Given armature resistance = 0.025 resistance = 60 and contact drop under each brush =

I V. Ignore armature reaction. (Elect. Machines (E-3) AMIE sec-c Winter 1991)

Solution. As Generator [Fig. 29.12 Load current.

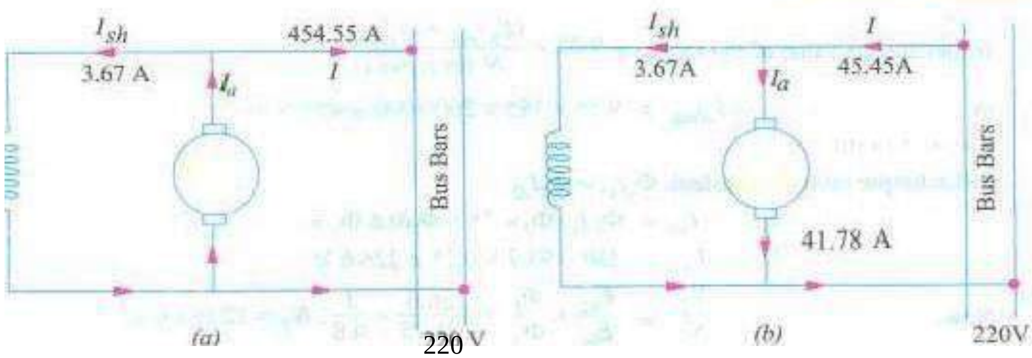


Fig. 29.12

As Motor [Fig. 29.12 (b)]

Input tine current = 100,000/220 = 45.45 A; I_{sh} current A: = 220/60 = 3.67 A

—45.45—3.67 —41.78 A; $I_a R_a = 41.78 \times 0.025 = 1.04$ V; E_{b2} -220— 1.04- 2X I —216.96 V

$$\frac{N_2}{N_1} = \frac{E_{b2}}{E_{b1}} \times \frac{\Phi_1}{\Phi_2}; \text{ since } \Phi_1 = \Phi_2$$

$$\frac{N_2}{300} = \frac{216.96}{233.45}; N_2 = 279 \text{ r.p.m.}$$

because is constant

Example 29.25. A dc. shunt machine generates 250, Von open circuit al 1000 r.p-m. Effective armature resistance is 0.5 Q. field resistance is 250 n, input to machine running as n motor on noload is 4 A at 250 V. Calculate speedofmachine as a motor taking 40 A at 250 V. Armatun• maction weakens field by (Electrical Machines-I, Univ.1987

Solution. Consider the case when the machine runs as motor on no-load.

Now, = 250/250 - A: Hence, G =-4— Em 250-0.5 x 3 2485 V

It is given that When armatureruns at 1000 it generates 250 V. When it generates V, it must be running at a speed = 1000 x 248.5/250 = 994 rpm.

ze, N_0 = 994 r.p.m_ Hence, 994

When n Loaded

$I_a = 40 - 1 = 39\ A;\ E_b = 250 - 39 \times 0.5 = 230.5\ V$ Also, $\Phi_0/\Phi = 1/0.96$

$$\frac{N}{E} = \frac{E_b}{E_{b0}} \quad \therefore \quad \frac{N}{994} = \frac{230.5}{248.5} \times \frac{1}{0.96}$$

N = 960 r.p.m.

Example 29.26.

A 250-V shunt motor giving /'1-92 kW al 1000 r.p.m. rakes an armature current of 75 A, The armature resistauce is 0.25 ohm and the load torque remains constant. If the flux is reduced by 20 percent Of its normal value before the speed changes, find rhe instantaneous value Of the armature current and the ,'orque. Determine the final Of the armature cur.'ent and speed.

œleet. Engg. AMIETE 'New Scheme) 1990 solution. Eh, - 250 – 75 × 0.25 = 231.25 V, as inFig.29.13.

75 A

When flux is reduced by 20%, the back e.m.f. is also reduced instantly by because speed remains constant due to inertiú of the heavy armature (Arl_ 29. Il).v

0.25 250

Instantaneous value ofback e.m.f. —231.25 0.8 = 185 v

= 260 A

Fig. 29.13

Instantaneous value of the torque = he torque = $9.55 \times \frac{(E_b)_{inst} \times (I_a)_{inst}}{N \ (in \ r.p.m.)}$

$(T_a)_{inst} = 9.55 \times 185 \times 260/1000 = 459$ N-m

Steady Conditiorrs

Since torque remains constant, nstant, $\Phi_1 I_{a1} = \Phi_2 I_{a2}$

$I_{a2} = \Phi_1 I_{a1}/\Phi_2 = 75 \times \Phi_1/0.8\ \Phi_1 = 93.7$ A

$E_{b2} = 250 - 93.7 \times 0.25 = 226.6$ V

Now. $\frac{N_2}{N_1} = \frac{E_{b2}}{E_{b1}} \times \frac{\Phi_1}{\Phi_2} = \frac{226.6}{231.25} \times \frac{1}{0.8}$; $N_2 =$ 1225 r.p.m.

Example 29.27. A 220-V, dc. shunt field resistance is 100 V. d.c. shunt motor takes 4 A at no-load when running at 700 r.p.m. The The resistance of armature at standstill gives gives a drop of 6 volts across armature terminals when 10 A were passed through it. Calculate (a) sp Calculate (a) speed on load armature terminals when 10 A were passed thmugh torque in At-m and (c) efficiency. The normal input Of the motor is 8 k W.

(Electrotechnies-n; MS. Univ. Baroda 1988 j

Solution. 200/100=2 A

EL. Power input

- 40-2-38 A; Ra=6/lO-O.6Ç2

— 198.8 V; 1772 V

; N —623.9 r.p.m.

700 198.8

T. 9.55 9.55 x 1772 x 38/623.9=103 N-m

(c) N. L. power input = 200 x 4-800 W; x 06—2.4 W

Constant losses - 800—2.4 = 797.6 W; FL. loss = 866.4 W

Total EL. losses 797.6 866.4 = 1664 W: output = 83M) - 1664 6336 W EL Motor efficiency = 0.792 or 79.2 %

Example 29.28. The input 23m V, dc. shunt motor is //kW. Calculate(a) the torque developed (h) the efficiency (e) the speed at this load. The particulars the molor areas follows :

No-load current = S A; No-load speed = r.p.m.

A rm. resistance = £2; shuntfield resistance = 110£2.

Solution. No—load input

No-load armature Cu loss

Constant losses When input is Il kW.

Input current Arm. Cu loss

Total loss

Output

Efficiency

(c) Back e.rn.f. at no-load
Back e.rn.f. at given load

Speed N

(Elect. Technology ;
Bombay University 1988' -
220x5= 1,100W:
$I_{sh} = 220/110 = 2A; I_{ao} = 5 -$

= 3^2 xo.5 W

= 1.100—4.5 =
1,095SW

= 11,0/220 = 50A ;
Armature current = 2 =
A

= 482 x 05 . - — 1,152 w;

= Arm. Cu loss + Constant
losses = 1 152 + 1095.5 =
2248 W

= 11.000- 2,248 — s, 752 w

8,752x

220-13 x 0.5) 218.5 V

= 196 v

196/218.5 =
r.p.m.

196 x 48

-87.1 N.m

Example 29.29. The armature circuit resistance a 18.65 kW 250- V series motor is O. / the brush voltage drop is 3V, and the seriesfield resistance is 0.05. When the motor takes 80 A, speed is

600 r.p.m. Calculate the speed when rhe current is *100 A.*

(Elect. Machines, A.M.I.E. sec. B, 1993b

Solution.

$$E_{b1} = 250 - 80\,(0.1 + 0.05) - 3 = 235\ \text{V}.$$
$$E_{b2} = 250 - 100\,(0.1 + 0.05) - 3 = 232\ \text{V}$$

Since 80/100

$$\Phi \propto I_a, \text{ hence, } \Phi_1 \propto 80,\ \Phi_2 \propto 100,\ \Phi_1/\Phi_2 = 8$$
$$\frac{N_2}{N_1} = \frac{E_{b2}}{E_{b1}} \times \frac{\Phi_1}{\Phi_2} \quad \text{or} \quad \frac{N_2}{600} = \frac{232}{235} \times \frac{80}{100}\ ;\ N_2 =$$

474 r.p.m.

Example A 220-volt d. c. series motor is running at a speed of 800 p.p.m. and draws 100 A_ Calculate at what speed rhe motor "'ill run when developing half the torq"e. Total msistance

(Elect. Machines ; A.M.I.E. Sec. B, 1991)

Of the armature andfield is 0.1 Ohm. Assume that the magnetic circuit is unsaturated.

Solution.

Since field is unsaturated,

$$\frac{N_2}{N_1} = \frac{E_{b2}}{E_{b1}} \times \frac{\Phi_1}{\Phi_2} = \frac{E_{b2}}{E_{b1}} \times \frac{I_{a1}}{I_{a2}} \qquad (\therefore\ \Phi \propto I_a)$$
$$\therefore\ T_a \propto \Phi I_a \propto I_a^2. \qquad (\therefore\ T_1 \propto I_{a1}^2 \text{ and } T_2 \propto I_{a2}^2)$$
$$T_2/T_1 = (I_{a2}/I_{a1})^2 \text{ or } 1/2 = (I_{a2}/I_{a1})^2\ ;\ I_{a1} = I_{a1}/\sqrt{2} = 70.7\ \text{A}$$
$$E_{b1} = 220 - 100 \times 0.1 = 210\ \text{V}\ ;\ E_{b2} = 220 - 0.1 \times 70.7 = 212.9\ \text{V}$$
$$\frac{N_2}{800} = \frac{212.9}{210} \times \frac{100}{70.7}\ ;\ N_2 = \mathbf{1147\ r.p.m.}$$

Motor Characteristics

The characteristic Curves Of a motor are those curves which show relationships between the following quantities.

I. and armature Current characteristic. It is nown as *electrical* Characteristic

2. Speed and armature current i.e. M/a characteristic.

3. Speed and torque i.e. characteristic. It is also known as mechanical chargcteristic. It can be found from (l) and (2) above.

While discussing •notor characteristics, the tòllowing two relations should always be kept in mind :

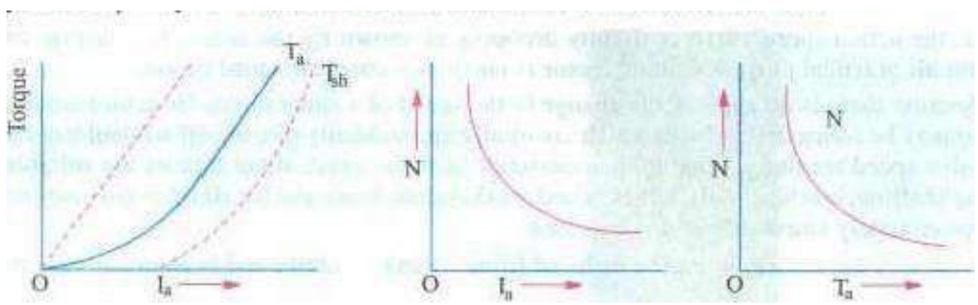
$T_a \propto \Phi I_a$ and N $\propto$

Characteristics of Series Motors

We have seen that T In this case, as field windings also earn' the armature current, rÞ up to the point of magnetic saturation, Hence. before saturation,

T and

At light loads. and hence is small. But as la increases, increases as the square of the current. Hence. Tall„ curve is a parabola as shown in Fig. 29.14. After saturation, is almost independent of hence To la only. So the characteristic becomes a straight line. The shaft torque T is less than torque due to Stray losses. It is shown dotted in the figure. so we conclude that (prior to magnetic saturation) on heavy loads, a series motor exerts a torque proportional to the square Of armature current. Hence, in cases where huge starting torque is required for accelerating heavy masses quickly as in hoists and electric trains etc.. series motors are used.



Fig, 29.14 Fig. 29.15 fig. 29.16

2. MI. Characteristics. Variations of SIEed can be deduced from the fGrrnula :

$$N \propto \frac{E_b}{\Phi}$$

Change in Eh, for various load currents is small and hence may be neglected for the time being. With increased la, also increases. Hence, speed varies inversely as armature current as shown in Fig. 29.15.

When load is heavy, ia is large. Hence, is low (this decreases Eb and allows more armature current to now). But when load current and hence la falls io a small value, speed becomes dangerously high. Hence. a series motor should never be started without some mechanical (not belt-driven) load on it otherwise it may develop excessive speed and get damaged due to heavy centrifugal forces so produced. It should be noted that series motor is a variable speed motor_

3. Or mechanical characteristic. It is found from above that when speed is high, torque is low and vice-verso. The relation between the two is as shown in Fig. 29.16.

Characteristics of Shunt Motors

I. Characteristic

Assuming CÞ to be practically constant (though at heavy loads. decreases somewhat due to increased armature reaction) we find that Ta .

Hence, the electrical characteristic as shown in Fig. 29.17. is practically a straight line through the origin. Shaft torque is shown dotted. Since a heavy starting load will need a heavy starting current, shunt motor should never be started on (heavy) load.

2. Characteristic

If is 'Issumcd constant, then N E,_ As Eh is also practically Constant. speed is, for most purposes. constant (Fig. 29.18).

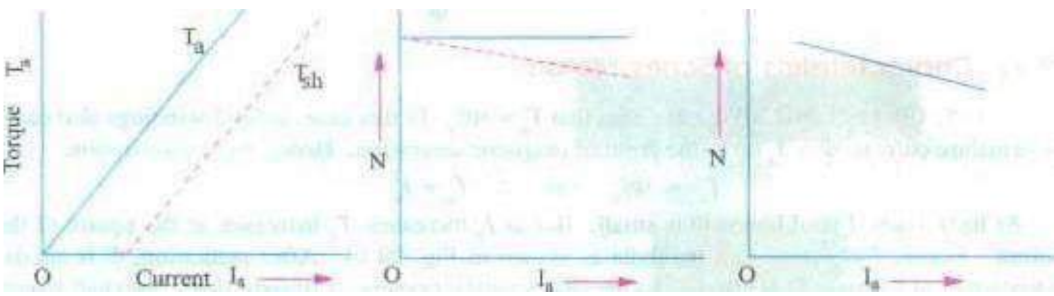


Fig. 29.17 Fig. 29.18 Fig. 29.19

But strictly speaking. both Eh and (Pde-crease with increasing load. However. Eb decreases slightly more than so that on the Whole, there is some decrease in speed. The drop varies from 5 to Of full-load speed. being dependent on saturation, armature reaction and brush position. Hence. the actual speed Curve is slightly drooping as shown by the dotted line in Fig. 29.18. But. for all practica] purposes, shunt motor is taken as a constant-speed motor.

Because there is no appreciable change in the speed of a shunt motor from no-load to fullload. it may be connected to loads which are totally and suddenly thrown off without any fear of excessive speed resulting. Due to the constancy of their speed. shunt motors are suitable for driving shafting. machine tools. lathes. wood-working machines and for all other purposes where an approximately constant speed is required.

3. N/T. Characteristic Can be deduced from and (2) above and is shown in Fig. 29.19_

29.15. Compound Motors

These motors have both series and shunt windings. If scries excitation helps the shunt excitation i.e. series flux is in the same direction (Fig. 29.20): then the tor is said to be cummulatively compounded- If on the other hand, series field opposes the shunt field, then the motor is said to be differentially compounded.

The characteristics of such motors lie in between those of shunt and series motors as shown in Fig. 29.21, Compound Motors

(a) Cumulative-compound Motors

Such machines are used where series characteristics are required and where, in addition,

the load is likely to be removed totally such as in some types of coal cutting machines or for driving heavy machine tools which have to take sudden cuts quite often. Due to shunt windings. speed will not become excessively high but due to series windings, it will be able to take heavy In conjunclion with fly-wheel ffunctioning as load equalizer), it is employed where there Fig. 29.20 are sudden temporary loads as in rolling mills. The fly-wheel supplies its stored kinetic energy when motor slows down due to sudden heavy load. And when due to the removal Of load motor speeds up, it gathers up its kinetic energy.

Compound-wound motors have greatest application With loads that require high starting torques or pulsating loads (because such motors smooth out the energy demand required of a pulsating load). They are used to drive electric shovels, metal-stamping machines. reciprocating pumps. hoists and compressors etc.

'b) Differential—compound Motors

Since series field opposes the shunt field; the flux is decreased as load is applied to the motor. This results in the motor speed remaining almost constant Or even increasing With increase in load (because, N Eb/(tD). Due to this reason, there is a decrexse in the rate at which the motor torque increases With load. Such motors are not in common use. But because they can be designed to give an accurately constant Speed under all conditions, they find limited application for experimental and research work.

One Of the biggest drawback Of such a motor is that due to weakening Of flux With increases in load. there is a tendency towards speed instability and motor running away unless designed properly.

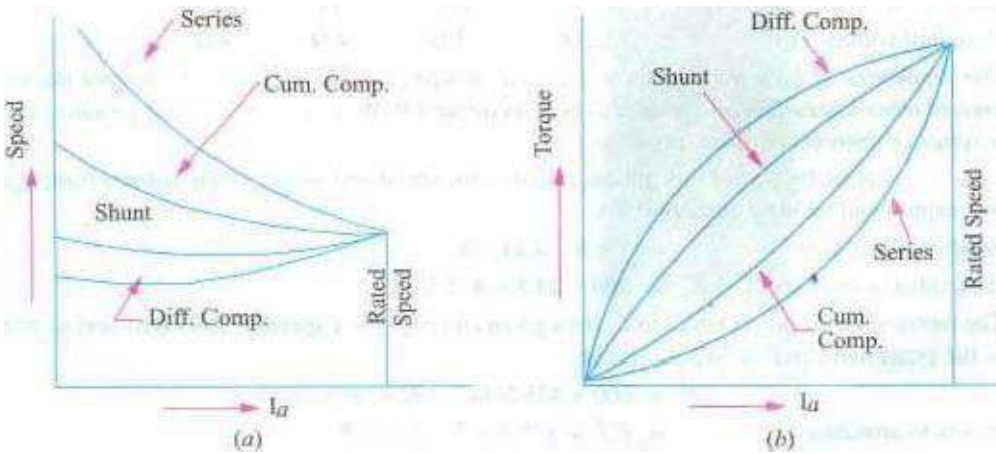


Fig.29Æi

Example 29—32. Thefollowing results were obtainedfmm a static torque a seriesmotor :

Current (A)	20	30	40	50
Torque (N • m)	128.8	230.5	349.8	46.2

Deduce the speed/tOrque curve for the machine When supplied at a constant voltage Of460 V. Resistance andfield winding is 0.5 Q. Ignore imn andfriction losses.

Solution. Taking the case When input current is

20 A, we have

Motor input = 460 x 20 = 9.200 W

Field and armature Cu loss

= 202 xos=200w

Ignoring iron and friction losses.

output = 9.200 − 200 = 9.000w

Now. Tah x 21t,V = Output in watts.

$128.8 \times 2\pi \times N =$ 9,000 0 100 200 300 400 soo N - 9,000/21tx 128.8 Torque (NW.m)

= 11.12 r.p.s. = 667r.p.m.

Similar calculations for Other values Of current are Fig. 29.22 tabulated below :

Current (A) 20 30 40 50

Input (W) loss (W)	9.200	13,800	18.400	23.000	1,250
Output (W)	9.200	13,350	11.600	21.850	
Speed (r.p.m.)	667	551	480	445	
Torque (N-tn)	128.8	2305	349.8	4692	

From these values, the speed/torque curve can be drawn as shown in Fig. 29.22.

Example Aýàn Which requires 8 (5.968 k W) at 700 r.p.m. is coupled directly to a d.c. series motor. Calculate the inpui tv the motor when the supply volraýe is5m V. assuming that power required for fan varies as the cube Of the speed. For the purpose Of obtaining the magnetisation characteristics. the motor was running as a self-excited generator al 600 r.p.m. and the relationship between the tenninal voltage and the load current was found ro be asfollows :

load current (A) 7 10.5 14 27.5 lèrminal voltage (V) 347 393 434 458

The resistance Of both the armature and field windings Of the motor is 3.5 Q and the core. friction and other losses may be assumed to be constant at 450 Wfor the speeds corresponding 10 the above range Ofcurrents at normal voltage. (London)

Solution. Let us. by way of illustration. calculate the speedand output when motor is running off a supply and taking a current of IA.

Series Voltage drop 7x3.5-24SV

Generated or back e.tn.f. Eh = —24.5 = 475—5 V

The motor speed is proponional to Eh for a given current. For a speed of 600 r.p.m. and a current

Of 7A. the generated e_m.f is 347 V. Hence,

N Ox475.5/347 -823 r.p.m.

Power to armature - -3.329W Output Armature —450 = 3,329 — 450 = 2.879 W = 2.879 kW power required by the fan at 823 r.p.m. is = 5.968 x $823^2/700^2$ = 9.498 kW

These calculations are repeated for the other values Of Current in the following table.

Input currrent (A)			14	27.5
Series drop (V)	24.5	36.7	49	96.4
Back e.m.f. (V)	475s	463.3	451	403.6

.F. at 600 r.p.m. (V)	347	393	434	458
Speed N (r.p.m.)	823	707	623	528
Armature power (W)	3329	4870	6310	1 1.100
Motor output W)	2.879	4.420	5.860	10.65
Power required by fan (kW)	9.698	6.146	4.222	2566

In Fig. 29.23 (i' the motor outputin kW and (ii' power required by fan in kW against input currentis plotted. Since motor output equals the inpul to fan. hence the intersection point of these curves gives the value of motor Input current under the given conditions.

(a)

Input current corresponding to intersection point = 12 A

∴ Motor input = 500 × 12 = **6,000 W**

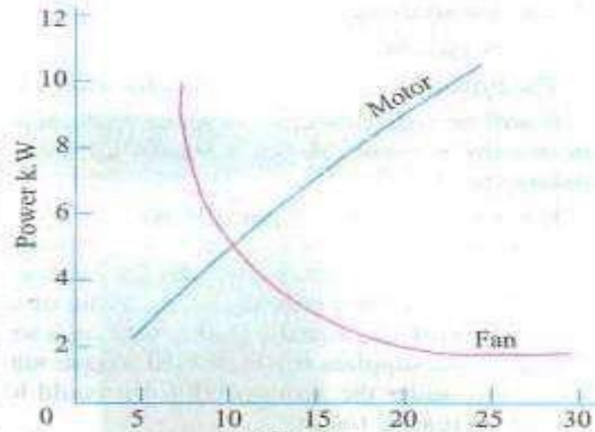


Fig. 29.23

29.16. Performance Curves

(a) Shunt Motor

In Fig. 29.24 the four essential characteristics of a shunt motor are shown *i.e.* torque, current speed and efficiency, each plotted as a function of motor output power. These are known as the ***performance curves*** of a motor.

It is seen that shunt motor has a definite no-load speed. Hence, it does not 'run away' when load is suddenly thrown off provided the field circuit remains closed. The drop in speed from no-load to full-load is small, hence this motor is usually referred to as ***constant speed*** motor. The speed for any load within the operating range of the motor can be readily obtained by varying the field current by means of a field rheostat.

The efficiency curve is usually of the same shape for all electric motors and generators. The shape of efficiency curve and the point of maximum efficiency can be varied considerably by the designer, though it is advantageous to have an efficiency curve which is farily flat, so that there is little change in efficiency between load and 25% overload and to have the maximum efficiency as near to the full load as possible.

It will be seen from the curves, that a certain value of current is required even when output is zero. The motor input under no-load conditions goes to meet the various losses occuring within the machine.

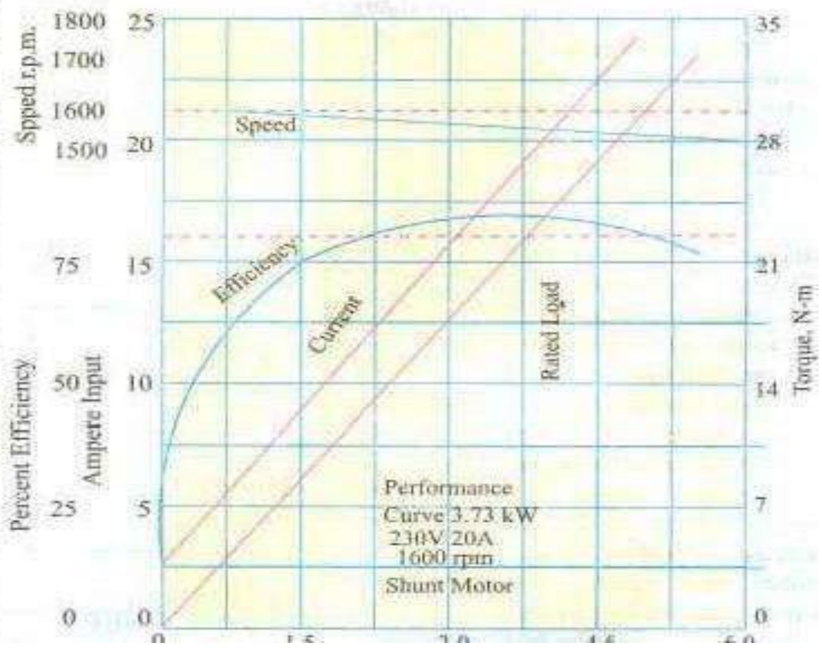


Fig. 29.24

As compared to other motors. a shunt motor is said to have a lower starting torque. this should not bc taken of mean that shunt motor is incapable Of starting a heavy load. Actually. it means that series and compound motors are capable of starting heavy loads with less excess of current inputs over normal values than the shunt motors and that consequently the depreciation on the motor

Will be relatively less. For example. if twice fuli load torque is required at Start, then shunt motor draws $I_a \propto \sqrt{T_a}$) whe twice the full-load current Iaor whereas series motor draws only $T_a \propto I_a^2$ or $I_a \propto$ approximately

one and a half times the fill/ load current

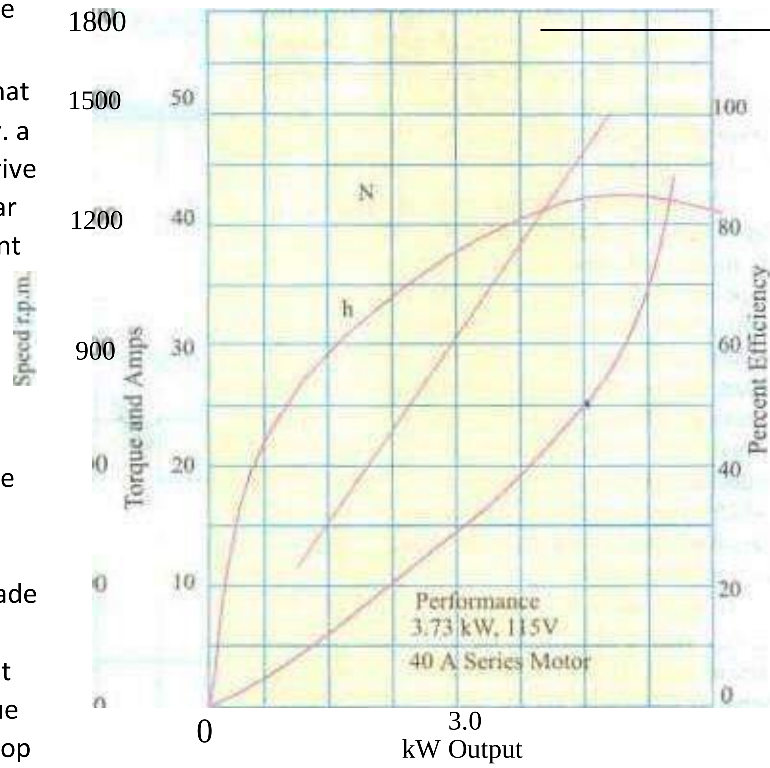
The shunt motor is widely used with loads that require essentially constant speed but where high starting torques are not needed- Such loads include centrifugal pumps. fans. winding reels conveyors and machine tools etc.

(b' Series Motor

The typical performance curves a series motor are shown in Fig. 29.25.

It will be seen that drop in speed with increased load is much more ptominent in series motor than in a shunt motor. Hence, a series motor is not suitable for applications requiring a Substantially constant svxed.

For a given curreng input, the siarting torque developed by a series motor is greater than that developed by a shunt motor. Hence. series motors are used where huge starting torques ure necessary i.e. for street cars, cranes, hoists and for electric-railway operation. In addition to the huge starling torque. there is another unique characteristic of series motors which makes them especially desirable for traction work i.e. When a load Comes on a series motor. it responds by decreasing its speed (and hence. E,) and supplies the increased torque with a small increase in On the other hand a shunt motor under the same conditions would hold its speed nearly constant and would supply the required increased torque With a large increase Of input current. Suppose that instead Of a series motor. a shunt motor is used to drive a Street car. When the ear ascends a grade, the shunt motor maintains the speed for car at approximately the Same value it had on the level ground. but the motor tends to take an excessive current. A series motor. however. automatically slows down on such a grade because Of increased current demand, and so it 600 develops more torque at reduced speed. The drop in speed permits the motor



to develop a large torque 300 With hut a moderate increase of power Hence. under the same load conditions. rating of the series motor would bc less than for a shunt motor,

Fig. 2925

29.17. Comparison Of Shunt and Series Motors

(a) Shunt Motors

The different characteristics have been discussed in Art.

29.14. It is clear that i") speed of a shunt motor is sufficiently constant.

(b) for the Same Current input, its starting torque is not a high as that of series motor. Hence. it is used.

(i) When the speed to be maintained approximately constant from NL to F.L i.e. for driving a line of shafting etc. ij When it is required to drive the load at various speeds. any onc speed being kept constant for relatively long period i.e. for individual driving of such machines as lathes. The shunt regulator enables the required speed to be obtained easily and economically.

Shunt Motors

Summary or Applications

Type of motor	Characterivtic•	Applicatio
Shunt	Approximately constant speed Adjustable speed Medium starting torque (Up to IS EL. torque)	For driving constant speed linc shafting Lathes Centrifugal pumps Machine tools Blowers and fans Reciprocating pumps
Series	Variable speed Adjustable variying Slxed High Starting torque	For traction work [e. Electric locomotives Rapid transit systems Trolley. ears etc. Cranes and hoists Conveyors
Comulalive Compound	Variable speed Adjustable varying speed High starting torque	For intermittent high torque loads For shears and punches Elevators Conveyors Heavy planers Heavy planers Rolling mills: Ice machines: Printing presses; Air compressors

(b) Series Motors

The operating characteristics have been discussed in Art 29.13. These motors have a relatively huge starting torques.

2. have good accelerating torque

3. have low speed high loads and dangerously high speed at low loads. Hence, such motors used

l. when a large starting torque required i.e. for driving hoists, cranes. trams etc.

2. when the motor can be directly coupled to a load such as a fan whose torque increases with

3. if constancy of speed is not essential, then. in fact. the decrease ofspeed with increase Of load has the advantage that the power absorbed by the motor does not increase as rapidly as the torque. For instance. when torque is doubled, the power approximately increases by about 50 to only

4. a series motor sh0uld not used where there is a possibility of the load decreasing to a Very small Thus, it should not be used for driving centrifugal pumps or for a belt-drive Of any kind.

Series Motors

Losses and Efficiency

The losses taking place in the motor are the same as in generators. These are (i) Coplrr losses (ii) Magnetic lcvsse_s and iiii• Mechanical losses,

The condition for maximum power developed by the motor is

$$I_a R_a = V/2 = E_b$$

The Condition for maximum efficiency is that armature Cu losses are equal to constant losses. (Art. 26.39).

Power Stages

The various stages of energy transformation in a motor and also the various losses occurring in it are shown in the now diagram of Fig. 29.26.

Overall or commercial efficiency —A' Electrical efficiency = —Ã , Mechanical efficiency

c

11m = -h

The efficiency curve for a motor is similar in shape to that for a generator (Art. 24ß5).

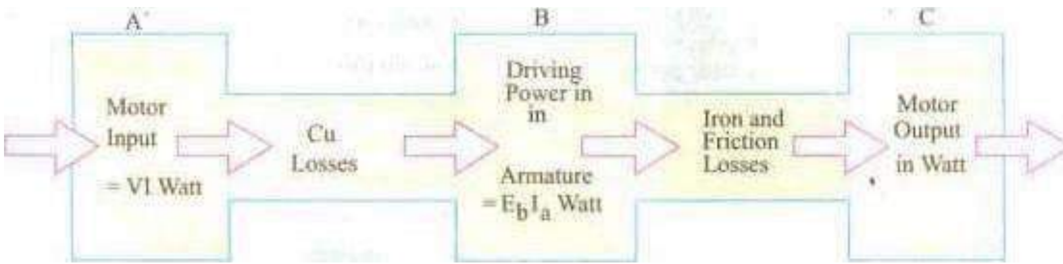


Fig.2926

It is seen that A — H = copper losses and B — C = iron and friction losses.

Example 29.34, One' ofthe r,vo similar 50th V' hunt machines A and B running lighl takes 3 A. When A is mechanically coupled to B. the input to A is 33 A with B unexcited and 4.5 A When B is separately-excited ro generate 500 V Calculate thefriction and windage loss and core loss ofeach machine. Machinery-I, Madras Univ.1985)

Solution. When running light. machine input is used to meet the following losses (i' armature Cu loss shunt Cu loss ' iii' iron loss and mechanical tosses i.e. friction and windagc losses. Obviously, these no-load losses fcu• each machine equal 500 X 3 W.

(a) With B unexcited

In this case. only mechanical losses take place in B. there being neither Cu loss nor iron-loss tecause B is unexcitcd_ Since machine A draws 05 A more currentFriction and windage loss of B 500 x 0.5 250 W

(b) With B excited

Inthis case, both iron losses as well as mechanical losses take place in machine B. NOW, machine

A draws, 4.5 — 3 = I .5 A more current.

Iron and mechanical losse• of B = IS X 500 750 W

Iron of B = 750 - 250 W

Example A 220 Shunt motor has an am-alary resistance Of0.2 ohm andfield resistance of The motor draws 5 A al V 500 r.p.m. at no load. Calculate the speed and shaft torque if the motor draws 52 A: at rated voltage. (Elect. Machines Nagpur Univ. 1993)

Solution.

210

1500 219.4

For finding the shaft torque. we will find the motor output when it draws a current of 52 A. First we will use the no-load data for finding the constant losses of the motor.

No load motor input 220 x 5 = W; Arm. Cu loss = 3^2 k W

∴ Constant or standing losses of the motor = 1100 1098

When oaded, arm. cu loss - 50^2 x 0.2 500 W

Hence. total motor losses = 1098 + 500 = 1598 W

Motor inputonload - 220 x 52 = 11,440 W; output — 11,440- 1598 = 9842 W

= 955 x outputhV = 9.55 x 9842/1436 N-m

Example 29—V'. 250 Vshunt motor on no runs 1000 and takes 5 amperes. Armature and shunt field resistances are 0.2 and 250 ohms respectively Calculate 'he speed when loaded taking current Of50 A. The armature r•ection weakens thefield by 3%.

(Elect. Engg.-l Nagpur Univ. 1993)

Sol

2402

1000 2492 0.97 q

Example 29.37. A 500 Vt}.c. shunt motor takes a current Of5 A On no-load The resislanees Of the armature andfield circuit are 0.22 ohm and 250 Ohm respectively. Find 'he efficiency when loaded and taking a current of 100 A (b) the pem.•enrage change of speed. State precisely the assumptions made. (Elect. Engg-l. MS. Univ. Baroda 1987)

Solution. No-Lond condition

$I_{sh} = 500/250 = 2$ A; $I_{a0} = 5 - 2 = 3$ A; $E_{b0} = 500 - (3 \times 0.22) = 499.34$ V

Arm. Cu loss $= 3^2 \times 0.22 = 2$ W; Motor input $= 500 \times 5 = 2500$ W

Constant losses = 2500-2=2498 W

It is assumed that these losses remain constant under all load conditions.

I .o:ad condition

(a) Motor current 100 A; Ia = $100 - 2 = 98$ A; $E_b = 500 - (98 \times 0.22)$ = 478.44 V

Example 29—39. A dc. shunt machine While running as generator develops a voltage Of 250 V at J000 on no-load. armature resistance Of 0.5£2 andfield resistance Of 250 When the machine runs us motor: inpul to it at no-load is 4.4 250 V. Calculate the speed and efficiency Ofthe machine it runs a motor taking 40 A at 250 Armature reaction weakens by (Electrical Technoloky. Aligarh Muslim Unis. '989'

Solution.

NOW. When running as a generator, the machine gives 250 V al 1000 r.p.m. If this machine was running as motor at 1000 r.p.m.. it will. obviously, have a back e.m.f. of 250 V produced in its armature. Hence N, = r.p.m. and Eh, 250 V.

When it runs as a motor. drawing 40 A. the back e.m.f. induced in its armature is Eb2 = $250-(40-1)\times 0.5=230.5$ V; Also $\Phi_2=0.96\ \Phi_1$, $N_2=?$

Using the above equation We have

$$\frac{230.5}{250}\times\frac{\Phi_1}{0.96\ \Phi_1};\ N_2=\mathbf{960\ r.p.m.}$$

1000

Emciencv

No-load input represents motor losses which consists of (a) armature Cu Ra Which is variable.

(b) constant losses W, which consists of jii shunt Cu loss (iib magnetic losses and 'iii' mechanical losses.

No-load input or total losses 250 x 4 W

Arm. cu loss = W, W, = = 995.5 W

When motor draws a line current of 40 A, its armature current is (40 — I = 39 A cu loss = 760.5 W: Total losses- 760.54 1756 W

Input 250x40= 10.000 W: output = 10.0 - 1756=8.244 W

- 8.244x IOW10.ooo=82..wrt

Example 29.40. The armature Winding Of a 4-poie•, 250 V dc. shunt motor is lap connected. There are 120 slots, each slot containing 8 conductors. The is 20 mWb and current taken by motor is 25 A. The resistance of armature andfield circuit are O. I and 125 respeclively. If the rotational losses amount 10 be 810 Wfind,

(i) gross torque (ii) useful torque and efficiency. (Elect. Machines Nagpur Univ. 1993)

Soil"ion. - (23 x 0.1) -247.7 V

20X IO-3 X960X,V 4

NOW, A 247.7 = 60 (4)— : N —773 r.p.nL

247.7X23

Gross torque or armature torque T. 9.55 9.55 x = 70.4 N - m

773

Rotational losses = 810 W: Total motor losses = 810 + 500 +53 = 1363 W

Motor input = 250 x 25 = 6250 W; Motor output - 6250 - 1363-4887 W

T_{sh} = 9.55 x outputiN = 9.55 x 4887/773 - 60-4 N-m

(iii) Efficiency = 4887/6250 = 0.782 = 782%

Example 29.11. A 20-hp (14.92 kWh 230-V, 1150-r.p.m. 4-pole, shun' motor has a total of 620 conductors arranged in two parallel paths and yielding an armature circuit resistance OfO. 2 Q

When it delivers rared power a' rated speed. it draws a line current Of 74.8 A and a field current Of 3 A. Calculate (i' theflu.x per pole the torque developed (iii)lhe rotational losses expressed as a percentage ofpower.

Solution.

(i) Now.

f Armature Torque,

(iii) Driving powerinarmature= Ei,la = 215.64* 71.8= 15.483 W

= 1 Outpul14,920 W; Rotational losses = 15.483—14.920= W VI = 230 × 74.8

(iv) Motor input- - 17.204 W: Total 17204- 14.920 W

Losses expressed as pcteentage Of power input = 2284/ 17,204 = O. 1 33 or 13.3%

Example 29.42. A 7.46 kW: 250• V shunt motor rakes a line current A When running light. Calculate the efficiency as a motor when deliveringfull load output, if the armature andjie/d resistame are 0.5 Q and 250£2 respectively. A,' output power will rhe efficiency be maximum ? Is it possible obtain this output from the machine ? MS. Univ. Baroda 1985)

Solution. When loaded lightly

Total motor input (Or total no-load losses) = 250 x 5 = 1.250 W

fieldCu10ss = 250* 1 =250 W;

∴ Iron losses and friction losses = 1250 — 250 — 8 = 992 W These losses would be assumed constant.

Let bc the full-load armature current, then armature input is = (250 x W EL output = 7.46 x 1000=7.460W The losses in the armature :

(i) Iron and friction losses = 992 W

(ii) Armature Cu loss = la: *0.05 w 2501. —_ 7,460 + 992 + I_a^2 ×0.5

O = 365 A

EL input curTent = 36.5+ I = 37.5 A ; Motor input = 250 x 37S W

EL output = 7,460 W

FL efficiency 100/250 x 375-79.6% NOW. efficiency is maximum When armature Cu loss equals Constant loss.

i.e. 1.242 Worla=49.84A

Armature input 250 x 49.84 12.460 W

Armature Cu loss 49.84^2 x 0.5 1242 W; Iron and friction losses = 992 W

Armatureoutput 10.226 w

Output power = 10,226 W = 10.226 kW

As the input current for maximum efficiency is beyond the full-load motor current, it is never realiscd in practice.

Input

Example 29.45. A 50-h.p. (373 kW), 460-V d.c. shunt motor running light takes a current of 4 A and runs ar a speed of 660 r.p.rn. The resistance of the armature circuit (including brushes) is 0.3 n and that ofthe shuntfield circuit 270 Q.

Determine when the motor is running atfull load

(ij rhe current input (ii the speed Determine the armature current at which efficiency is gnat-imam. Ignore the effect ofarmature reaction.1991)

Solution. 460/270= 1.7

When running light

I_a= 4— 1.7 = 2.3 A; Armature Cu loss —2.3^2 *0.3 1.5 W (negligible)

No-load armature input = 460 x 2.3 = 1 ,058 W

As armature Cu loss is negligible, hence 1,058 W represents iron. friction and windage losses which will assumed to be constant.

Let full-load armature input current be / Then

Armature input = I W; Armature Cu loss x 0.3 W

= 38.358=0 la = 88.5 A

fij Current input= 88.5+ 1.7=90.2A

= 660 x 433.5/459a=624r.p.m.

For maximum efficiency, 10^2 Ra constant losses (Art. 24.37) exo.3 = 1,841

Tutorial Problems 29.3

A 4-POle 250—V. motor a armature With

(a) the torque (b) the sreed

(b) the outpat torque and (d) the efficiency. if the motor current is 50 A The value Of flux per pole under these conditions is 22 mWh and the corresponding iron. friction and wirxtagc losses total 810 W. Armature resistancc O. 19 field resistance 0.14 Q.

Ha) 173.5 N-rn(b) r.p.m. (c) N.m (d) 86.9%]

2. no-load, a shunt tyotor takes 5 A at 250 V. resistances of the field and armature circuits are 250 Q and O. I respectively. Calculate the output power andefficiency of motor when the total supply current is 81 A at the same supply voltage. State any assumptions made.

91 % . It is that windage. friction and eddy current Josses are independent of the current and speedl

SINGLE PHASE TRANSFORMER

. Working Principle Of a Transformer

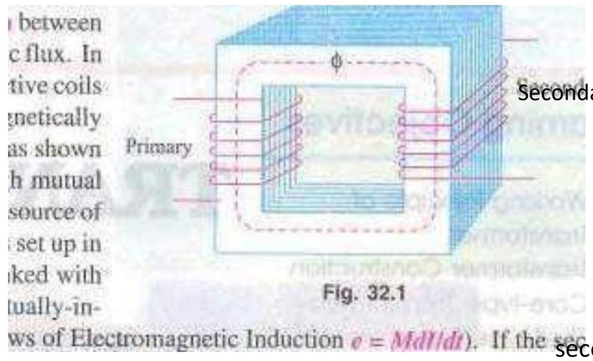


Fig. 32.1

Secondary

second

A transformer is a static (or stationary) piece ofapparatus by means of which electric power in onecircuit is transformed into electric power of the same frequency in another circuit. It can raise or lower the voltage in a circuit hut With a corresponding decrease or increase in curmt. The physiCal Laminated Core basisofa transformeris mutual induction between two cimuits linked by a common magnetic flux. In its simplest form, it consists Of two inductive which are electrically separated bat magnetically linked through a path of low reluctance as in Fig. 32.1. The two coils possess high mutual inductance. If One coil is Connected to a alternating voltage, an altemating flux is set up in the laminated core. most of which is linked the other coil in which it produces mutually-induced e.m,f.facCording to Faraday's Laws Of Electromagnetic coil circuit is closed, a current flows in it and so electric energy is transrewed (entirely magnetically) from the first to the second coil. The first coil, in which electric energy is fed from thea.c. supply mains, is calledprimary winding and the other from which energy is drawnout, is called secondary winding. In brief. a tKånsformer is a device that

1. transfers electric power from one circuit to another

2: it does so without a change of frequency

3. itaccomplishes this by electromagnetic induction and

4. where the two electric circuits arein mutual inductive influence of eachother.

32.2. Transformer Construction Ironcore

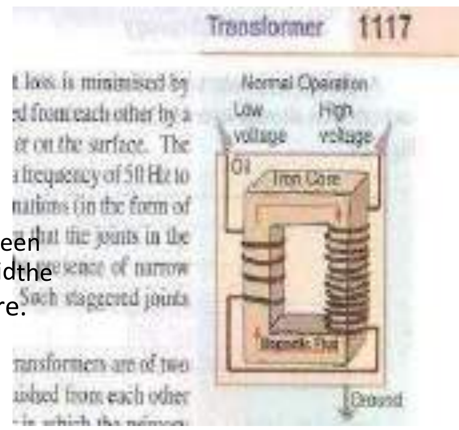
The simple elements of a transformer consist of two coils having mutual inductance and a laminated steel core. The primary secondary two coils arc insulated from each other and the steel core. Other necessary parts are :

some suitable container for assembled core 110/120 220/240; and windings ; suitable medium for Volts insulating the core and its windings from its container suitable bushings (either of porcelain, oil-filled or capacitor-type) for ondary insulating and bringing out the terminals of windings from "0/120

types ofPrinciple ot transformer

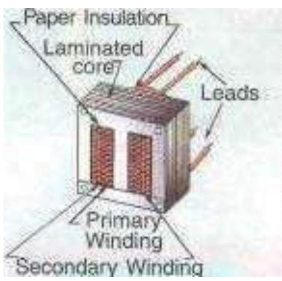
the core is constructed of transformer sheet steel laminations assembled to provide continuous magnetic path With a minimum of air-gap included. The steel used is of high silicon content, sometimes heat treated

Fig. 32.2 to produce ahigh permeability and a low hysteresis loss at the



usual operating flux densities. The eddy current loss laminating the core, the laminationsbeing insulated from each other by a light coat of cote-plate vanlish or by an oxide layer on the surface. The thickness of laminations valieS from 0.35 mm for a frequency of 0.5 mm for a frequency Of 25 Hr.. The core laminations (in the strips) are joined as shown in Fig. 32.2. It alternate layers are staggered in order to gaps right through the cross-section of the are said to be 'imbricated'.

isseen avoidthe core.



Constructional/y, the general types, distinguished from each other merely by the manner in which the primary

Core-type transformer and secondary coils are placed around the laminated core. The two types ale known as (i) core-type and (ii) shentype. Another recent development is spiral-core-or wo«nd-coretype, lhc trade name beingspirakore transformer,

In the so-called core type transformers, the windings surround a ary Winding considerable par/ ofthe core whereas in shell-type transformers. the core

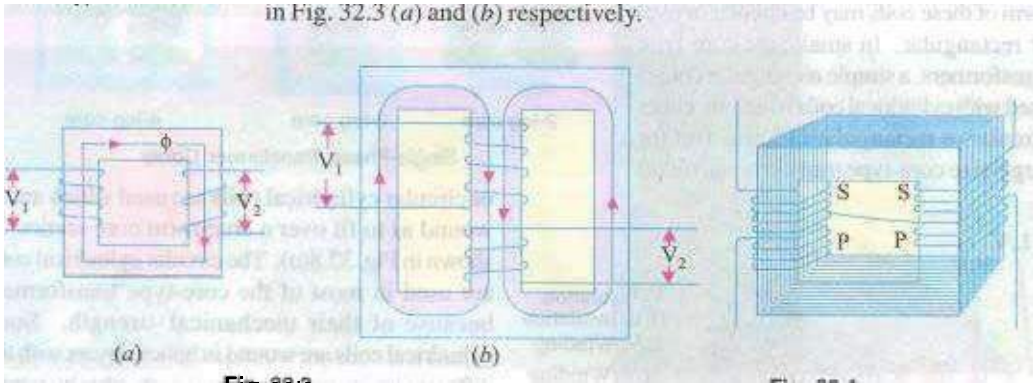
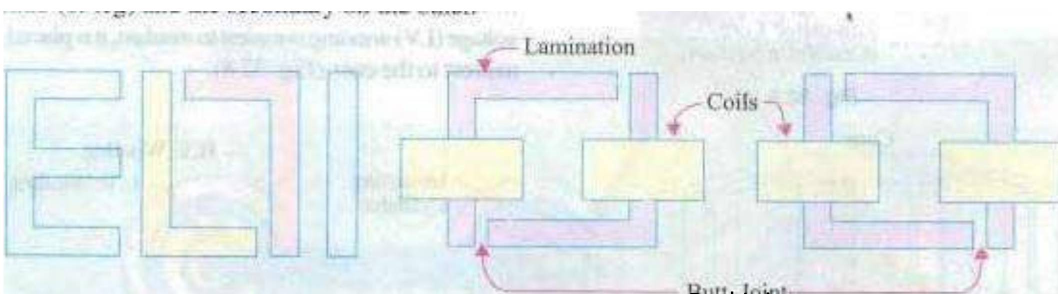


Fig.32'3

Shell-Type transformer surrounds a considerableportion ofrhe windings asshownschematically in Fig. 32.3 (a) and (b) respectively

In thesimplified diagram for the core type transformers [Fig. 32.3 the primary and secondary winding are shown located on the opposite legs (or limbs) of the core. but in actual construction, these are always interleaved to reduce leakage llux_ As shown in Fig. 32.4, halfthe primary and halfthe secondary winding have been placed side byside or concentrically on each limb, not primary on one limb (or leg) and the secondary On the other„



Butl Joi

Fig. 32.5 Fig. 32.6

In both core and shell-type transformers, the individual laminations are cut in the form of long strips of Vs, Es and rs as shown in Fig. 32.5. The assembly of the complete core for the two types of transformers is shown in Fig. 32.6 and Fig. 32.7.

As saidabove, in order to avoid high reluctance at thejoints where the laminationsare buttedagainst each other, the alternate layersare stackeddifferently to eliminatethesejointsas shown in Fig. 32.6and

Fig. 32.7:

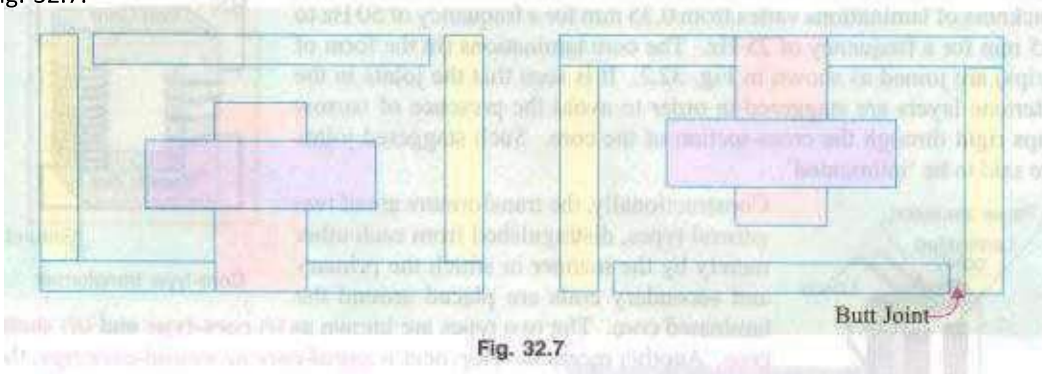


Fig. 32.7

Because oflaminations and insulation, the net oreffective area is reduced. due allowance for which has to be made (Ex. 32.6). 'tis found that. in general, the reduction in coresectional areadueto the presence ofpaper, surface oxide etc. is of the order of 10% approximately.

As pointed out above. rectangularcores With ectangular cylindrical coils can be used forsmall-size cott-ty•pc transformers as shown in fig. 32.9 (a) but for large•si7Æd transformers, it becomes wasteful to use rectangular cylindrical coils and so circular Cylindrical coils are preferred. For such purposes. square cores may be used as shown in Fig. 329 (b) where circles represent the tubular former carry ing the coils. Obviously. a considerable amount of useful space is still wasted. A common improvement on square core is to employ cruciform Core as in Fig. 32.9 (C) which demands, at least, two sizes Of core strips. For very large transformers. further core-stepping is done as in fig. 32.9 (d) where at least three sizes of core plates are necessary. not only gives high space factor but also results in reduced length ofthe mean turn and the consequent R loss. Three stepped core is the one most commonly used although more Steps may be forveo' large transformers as in fig. 32.9 From the geometry Offig. 329, it can shown that maximum gross core section for Fig. 32.9 (b) is 0.5 and for Fig. 32.9 (c) it is 0.616 where d is the diameterofttE cylindrical coil.

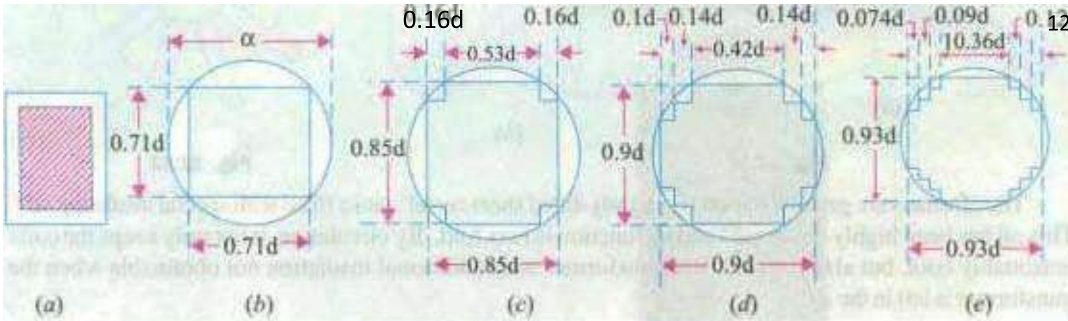


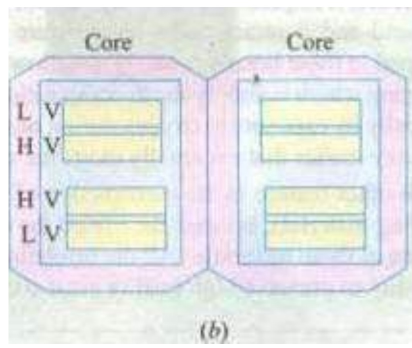
Fig. 32.9

32.4. Shell-type Transformers

In these case also, the coils are form-would but are multi-layer disc type usually wound in the

In these case also, the coils are form-would but are multi-layer disc form of pancakes. The different layers of such multi-layer discs are insulated from each other by pavrr. The complete winding consists of stacked discs with insulation space between the coils—the spaces forming horizontal cooling and insulating ducts. A shell-type transformer may have a simple

gular form as shown in Fig. 32.10



rectangular form as shown in Fig. 32. or it may have distributed form as shown in Fig. 32. I I.

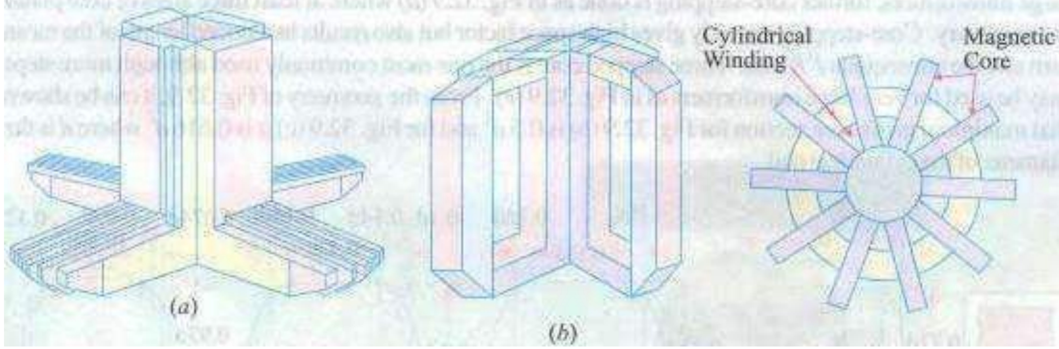
Fig. 32.10

A very commonly-used shell-type transformer is the one known as Berry Transformer—so called after the name of its designer and is cylindrical in form. The transformer core consists Of laminations arranged in groups which radiate out from the centre as shown in section in Fig. 32.12.

It may be poi ated out that cores and coils Oftransformers must be provided With rigidmechanical bracing in Olderto prevent movement and possible insulation damage. bracing reduces vibrationand the objectionable noise—a hummingsound-duringoperation.

The spiral-core transformer employs the newest development in construction. The core is asembled of a continuous strip or ribbon of transformer steel wound in the form of a circular or elliptical cylinder. Such construction allows the core flux to follow the grain Of the iron. Cold-rolled steel Of high silicon content enables the designer to use considerably higher operating flux densities With lower loss per kg. The use of higher flux density œduees the weight per WA. Hence, the advantages ofsuch construction are

ure (i) a relatively more rigid core (ii) lesser weight and size per kVA rating (iii) lower iron losses at higher operating flux densities and (iv) lower cost of manufacture.



li

a relatively more rigid (ii) lesser weight and size per kVA rating (iiö lower iron losses athigher

Fig. 32.11 Fig. 32.12

Transformers ale generally housed in tightly-fitted sheet-metal ; tanks filled with special insulating This oil has been highly developed and its function is two-fold. By circulation, it not only keeps the coils reasonably cool , but also provides the transformer with additional insulation not obtainable when the transformer is-left in the air.

In cases Where a smooth tank surface does not proyide sufficient cooling area, the sides Ofthe tank are corrugated or provided with radiators mountedon the sides. Good transformer oil should he absolutely free from alkalies, sulphur and particularly from moisture. The presence of even an extremely small percentage ofmoisture in the oil is highly detrimental from the insulation viewpoint because it lowers the dielectric strength Of the Oil considerably. The importanceofavoiding moisture in the transformer Oil is clear from the fact an addition Of 8 parts of Water in I reduces the insulating quality Of the Oil to a value generally recogniud as below standard. Hence, the tanks are sealed air-tight in smaller units. In the case of large-sized transformers where complete air-tight construction is impossible, chambers known as breathers are provided to permit the Oil inside the tank to expand and contract as its temperature increases or decreases. The atmospheric moisture is entrapped in these breathers and is not allowed to pass on to the oil. Another thing 10 avoid in the oil is sledging which is simply the decomposition of oil with long and continued use. Sledging is caused principally by exposure to oxygen•during heating and results in the formation Of large deposits ofdark and heavy matter that eventually clogs the cooling ducts the transformer.

NO Other feature in the construction Of a transformer is given more attention and care than the insulating materials. because the life on the unit almost solely depends on the quality. durability and handling of these materials. All the insulating materials are selected on the basis of their high quality and ability to preserve high qualityeven after many years of normal use.

Insteadofnatural mineral oil. now-a-days synthetic ilLsulating fluids known as ASKAREIAS (trade name) are used. They arc non-inflammable and, ihe influence of an electric arc, do not decompose to produce inflammableeases. One such fluid commercially known as PYROCLOR is being eOensively used because it possesses remarkablestability as adielectric and even after long shows no deterioration through sledging. oxidation,acidormoisture formation. Unlike mineral Oil, iUshowsmo rapid burning.

All the transformer leadsarebroughtoutoftheircases through suitable bushingS. There are many designs Of the-SC, their-size and constnlction depending on the voltage Of the leads. For moderate voltages, porcelain bushings are Lsed to insulate the leadsas they come out through the tank. In general, they look almost like the insulators used on the transmission lines. Inhigh voltage installations, Oil-filled orcapacitor— type bushings areemployed.

The choice ofcore or shell-type construction is usually determined by cost, because similar characteristics can be obtained withboth types. For very high-voltage transformers or formultiwinding design, shelltype constmction is preferred by many manufacturers. In this type. usually the mean length ofcoil turn is longer than in a comparable core-type design. Both core and shell forms areused and the selection is decided by many such as voltage rafing, kVA rating, weight. insulation stress, heat distributionetc.

Another means of classifying the transformers is according to the typeofcooling employed. The following types are in common use :

(a) oil-Aled self-cooled (b) oil-filled Wafer-cooled (c' áir-biast type

Small and medium size distribution transformers-so called because of their use on distribution systems as distinguished from line transmission—are of type The assembled windings and cores of such transformers are mounted in a welded, oil.tight steel tank provided with steel cover. After putting the core at its proper place, the tank is filled with purified, high quality insulating oil. The oil senes to convey the heat from the core and the windings to the case from where it is radiated the surroundings. For small size, the tanks are usually smooth-surfaced, but for largersizes. the cases are frequently corrugated or fluted to get greater heat radiation area without increasing the Cubical capacity Of the tank. Still larger sizes are provided With radiators or pipes.

Construction Of very large self-cooled transformers is expensive, a more economical form Of Construction for such large transformers is provided in the oil-immersed, water-cooled type. As before, the windings and the cote are immersed in the Oil, but there is mounted near the Surface of oil, a coolingcoil thmughâh'hichcold Water is kept circulating. The heat is carried awayby this water. The largest transformers such as those used With high-voltage transmission lines; are constructed in this

Oil-filled transformers are bail t for outdoor duty and as these require no housing Other than their own. a great saving is thereby eflCcted. These transformers require only periodic inspection.

For voltages below 25.000 V. transformers can bc built for cooling by means of an air-blast. The transformer is not immersed in oil. bat is housed in a thin sheet-metal box open at both ends through which air is blown from the bottom to the top hy means of a fan or blower.

Elementary Theory of an Ideal Transformer

An ideal transformer is one which has no losses i.e. its windingshave noohmic resistance. there is no magnetic leakage and hence which has no 1^2R and core losses. In other words. an ideal transformerconsists oftwo purely inductive wound on a loss-free core. It may, however, be notedthai iv is impoçsible

10 realize such a transformerin practice, yet ror cani•enience, will start with such a former

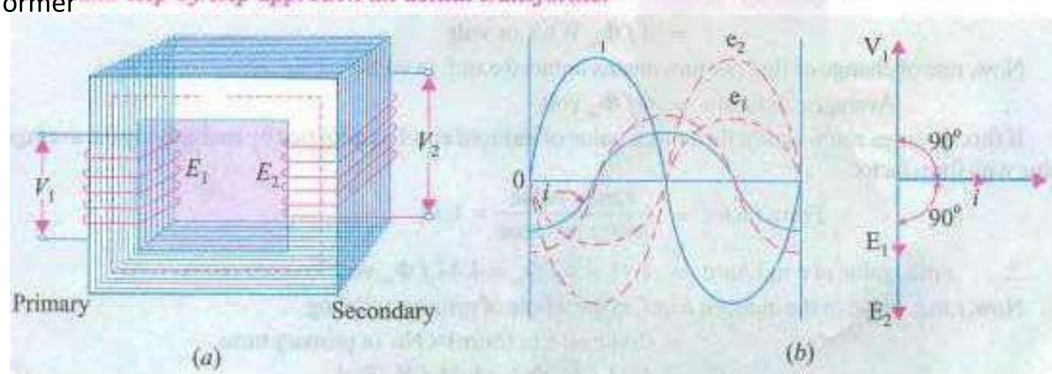
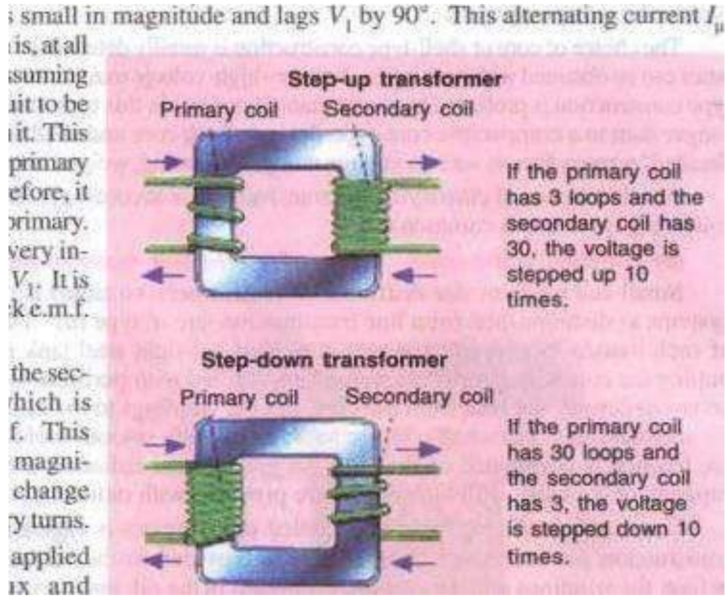


Fig. 32.13

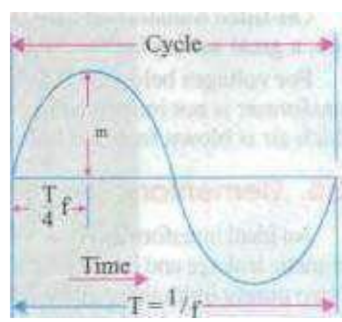
Consider an ideal transformer [Fig. 32.13 whose secondary is open and whose primaryis connected to sinusoidal alternating voltage VI. This potential difference causes an alternating curtent to flow in theprimary. Since the primary coil is purely inductive and there is no output (seconda/y being open) the primary draws the magnetising current only. The function Of this current is merely to magnetise the core. it is small in magnitåde and lags VI by 90•. produces an alternating flux Owhich is. at all times, pmportional to the cun-ent OLssuming permeability of the magnetic circuit to be constant) hence. is in phase with it. This changing flux islinked bothwith the and the secondary windings. Therefore, it produces self-induced e.m.f, in the primary This self-induced e.m.f. Et is. at every inStant, equal to and in to VI. It is also known as Counter e.m.f. or back e. ofthe primary.



Similarly, there is pnxiuced in the secondary an induced e.tn.f. E2 which is known as mutually induced c-m.f. This e.m.f. is antiphase with V, and its magnitude is proportional to the rate of change Of flux and the numberof secondary turns

The instantaneous values Of applied voltage. induced e.m.fs. flux and magnetising current are shown by sinu- Step-up transformer soidal wavesinFig. 32.13 fig. 32.13

(c) shows the Vectorial repœsentation Ofthe effective values Of the above quantities.



E.M.F. Equation Of a Transformer

Let NI = No. Of turns in primary

Nk = No. ofturns in secondary

= Maximum flux in core in w&rs

f =Frequency of a.c, input in Hz

As shown in fig. 32.14, flux increases fromits zero value to maximum value 4'nin one quarter of the cycle i.e. in 1/4/second.

$= \frac{\Phi_m}{1/4f}$

$4f\Phi_m$ Wb/s

∴Average rate of change of flux =

Fig. 3214

= or volt

Now. rate ofchangeofflux perturn means induced e_m-f_ involts_

Avetagee.rn.f./turn = 4fcbm volt

If flux varies sinusoidally then r.m.s. value ofinduced e.m.f. isobtained by multiplying theaverage valuewith formfactor. rm. S. value

Formfactor =-1.11

average value

∴rms. valueofe.m.f]turn = 1.11 x 4 4.44 volt

Now. r.ms. valueofthe induced e.m.f. in the whole ofprimary winding

= (inducede.tn.fjturn)xNo. ofprimarytums

Hence. currents aræin the inverse ratioof the (voltage) transformation ratio.

Example 32.1. The ^{2}maximumflux density:tn die core ofa 250/3000-Võlts,SO-Hzsingle-Þhase transformer is 1.2 Wb,4n. Ifthe e.m.j: per furn is 8 volt, determine (i) primary and secondary turns areg core.

(ElectricalEngg.-l, Nagpnr Univ. 1991)

Solution. (i} E = xe.m.f.induced/turn

(ii) We may use = — 4.44 Bm A

3000 4.44x 50 x 375 xL2xA; A = 0.03m2.

Example 32.2. The of a IOO-kVA. 11000/550 V: _50-HZ„ I -ph, Copé type transformer has a cross-section of20cmX 20em. Find the numberofH. V,and L turns perphase and the e,m.f: per turn if rhe coredensiry is nor 'o exceed Tesla. Assume a stacking factor Of0.9' What will happen if its primary voltage increased bv on no-load ?

(Elect. Machines, AM.LE, Sec. By 1991)

Solution. 1.3

11,000 = x 1060

550 = xL3 X 0.036: N2=53

= 1060=53

e.m.f.}tum = 10.4 V or 550/53= 104 v

Keeping supply frequency constant, ifprimary voltage is increased by 10%, magnetising current will incœase by much more than 10%. However. due to saturation, flux density Will increase only and sa will the eddy current and hysteresis losses.

Example 323. A single-phase transformer has 400 primary and 1000 secondary turns. The net cross-sectional area ofthe core is 60 cm^2 . Iftheprimary winding be connected a 50-Hz supply ar 520 V. calculate (i) the peak value density in the core (ìi) the voltage induced in

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Solution. $K = N_2/N_1 = 1000/400 = 2.5$

(*i*) $E_2/E_1 = K \quad \therefore \quad E_2 = KE_1 = 2.5 \times 520 = 1300$ V

(*ii*) $E_1 = 4.44\, f N_1\, B_m A$

or $520 = 4.44 \times 50 \times 400 \times B_m \times (60 \times 10^{-4}) \therefore B_m = 0.976\ Wb/m^2$

the secondary Winding. (Elect. Engg-E. rune Unh'. 1989) Example 32.4, '4 25-kVA transformer has 500 turns on the primary and 50 turns on the second. ary winding. The primary is connected to 3000-1' SO-Hz supply. Find the full-load primary and secondary currenfs, the secondary e.nt.f: and the maximum flux in the core. Neglect leakage drops and no-load primary cgrrenÇ (Elect. & Elect.mnic Engg„ Madras Univ. 1985)

Solution. K = N2/Nt-50/500-1/10

Now, full-load

e.m.f.per turn on primary side = 30/500—6 Secondary e.m.f.

r operates With a number ofH.V. and L V.

(iii) full load H. V. and L phase-currents,

SolutionMaximumvalue offlux has twen given as 0.05 Wb.

(*ii*) e.m.f. per turn 4.44 f

= 4.44 x 50 x 0.05 = 11.1 volts

(*i*)Calculations fornumtvofturns on two sides

Voltage phase on delta-connected primary winding 11000 volts

Voltage per phase on star-connected secondary winding 550/1.732 = 317.5 volts T1 = numberofturnson primary, per phase voltage perphase}e.m.f. turn

T. numberof turns on secondary. per phase voltage per phase/e.m.f. per turn 317.5/11.1 - 28.6

Note : Generally, Low-voltage-turns calculated first. the figure is rounded off to next higlEr even integer. _In this case. it be 30. Theo. number Of turns on primary side is calculated by turns-ratio.

In this case,

This. however, reduces the flux and results into less saturation. This, in fact, is an elementary aspect in Design-calculations for transformers. (Explanation is added here only to overcome a doubt whether a fraction is acceptable as a number ofL.V. turns).

(ii) Full load and L.V. phase currents :

Output per phase =

H.V. phase-current

L.V. phase-current =

Example 32.6. A single{phase tranžformer has 500 turns in the primary and turns in the secondary. The cross-sectional area Ofthe core is 80 sq. cm. (fthe primary windingì.ç connected to a 50 H: supply at500 V. calculate (i) Peakflux-density, and lii) Voltage induced in the secondary.

(Bharathinr University November 199T

, Solution. From the

(i) Peakfluxdensity, 20

(ii) Wtageinducedin secondary is from tmnsformationratioorturnsratio

.m.f. equation for transformer,

$$500 = 4.44 \times 50 \times \phi_m \times 500$$

$$\phi_m = 1/222 \text{ Wb}$$

$$B_m = \phi_m / (80 \times 10^{-4}) = 0.563 \text{ wb/m}$$

obtained

$$\frac{V_2}{V_1} = \frac{N_2}{N_1}$$

$$V_2 = 500 \times 1200/5$$

V2 — 500k12CW500= 1200volts

Example 32.7. A 25 kVA. single•phase trttnSfOñ7ier41ås 250 rurnS on primàrÿ and 40 thè secondary winding. The primary is connected to 50 H: mains. Calculate (i) primary and Secondary currents on"'Il-load, (ii)$ecendary e.m.f.. (ìii) maximumflux in the core.

(Bharathiar Univ. April 1998)

Solution.= Secondary voltage rating, =secondary e.m.f.'

$$\text{If } \frac{V_2}{V_2}$$

1500 — 250 , giving V2 = 240 volts

(in Primary current 25000/1500-16.67amp Secondary curent 25000/240— 104.2amp

(iiib If0mis the maximum core-flux in Wb;

1500 = 4.44 x givingÇ=O.027Wbor27mWb

Example 32.8. A asingle-phase, 50 "z, core-type transformer has Square cores of20 cm side. Permissible maximumflux-density is" Wb/m^2. Calculate number of 'urns per Limb on•he High and larw•volroge sides for u 3000/220 V ratio. Manonmnnium Sundaranar cniv. April 1998)

Solution. E-MF. equation gives the number of turns required on the two sides. We shall first calculate the L.V.-turns, round the figure off to the next higher even number, so that given maximum

flux density is not exceeded. With the corrected number of L.V. turns, calculate H.V.-turns by transformation ratio. Further, there are two Limbs. Each Limb accommodates half-L V. and halfH.V]

Winding from the view-point of reducing leakage reactance.

Starting with calculation for LV. tums, Ty

4.44x50xt(20x20x Tz=220

= 220/8.0=24.77

Select

= 26 x 3000/220 = 354, selecting the nearest even integer.

of H.V turns on each Limb = /

Number OfLV. turns on each Limb= 13

Transformer with Losses but no Magnetic Leakage

We will consider two cases (i) when such a transformer is On no load and (i" when it is loaded.

Transformer on No-load

In the above discussion. we assumed ideal transformer i.e. One in which there Were no losses and copperlosses. But practical conditions require that certain modifications be made in the foregoing theory. When an actual transformer is put on load. there iviroiilošsin the core andcopperlossin the Windings (both primary and secondary) and these losses arenot entirely negligible.

Even when the transformer is on no-load, the primary input current is not wholly reactive. Theprimary input current under nozload conditions hastosupply (i) iron losses in the core i.e. hysteresis lossand eddy current loss and (ri) a very small copper le8s in primary (there being no Cu loss in seconda/y as it is open). Hence, the ncAoad primary input Current at 90^0 behind VI but lags it by an angle $\phi_0 <$ 90^0 . No-load input power

Wo = Vilo cos ϕ_0

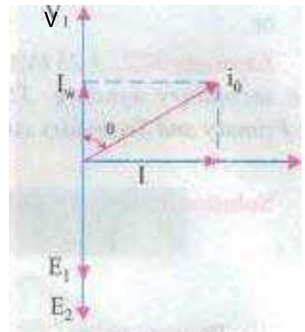


Fig. 32.16

Where cos is primary power factor underm-load Conditions. No-load Condition pf an actual transformer is shown vectorially in Fig. 32.16.

As seen from Fig. 32.16, primary current 10 has two components :

(i) One in phase With V,. This is known as active or working or iron loss component Iw because it mainly supplies the iron loss plus small quantity of primary Cu loss,

I_w =10 COSϕ_0

(ii) The Other component is in quadrature With VI and is known as magnerising Component I because its function is to sustain the alternating flux in the core. It is wattless.

losin4)o

Obviously, 10 the vector sum of [wand I . hence The following should noted carefully :

l. The no-load primary current Ibis small as compared to the full-load primary current. It is about 1 per cent of the full-load current.

2. Owing to the fáét that the permeabilityofthe cofe varies with the instantaneous value of the exciting current, the wave of the exciting or magnetising current is not truly sinusoidal. As such it should not be represented by a vector because only sinusoidally varying quantities are represented by rotating vectors. But, in practice, it makes no appreciable difference.

3. As 10 is very small, the no-load primalY Cu loss is negligibly small which means that no-load primary input is practically equal t" the iron loss in the transformer.

4. As it is principally the core-loss Whichis responsible for shiftin the known as hysteresis' angle advance.

Example 32.9. (a) A 2, 200/200-1' transformer draws a no- load primary èurrent Of0.6 A and absorbs 400 watts. Find the magnetising and iron loss currents.

(h) A 2200/250-1" transformer takes A at a p.f Of 0.3 on open circuit. Find magnetising and working components of no-load primary current.

Solution. Iron-loss current

no-load input in watts 400 -- = 0.182 A primary voltage 200

Magnetising component — 0.182)? 0.672 A

The two components are shown in fig. 29 5.

Example 32.10. A Single-phase transfoñnèr/tas 500 turns on the primary andJ4tJturns on the secondary winding. The•neon length Of rhe magnetic path in rhe iron core is 150 cm and the joints are quivalent'o an air-gap ofO. When a p.d. V is applied to the primary maximum flux density is 1.2 Wb/m^2. Calculate (a/ the cross-sectional areaof the core (b) no-load secondary voltage (c) rhe no-load current drawn by the primary (d) power factor no-load. Given that AT/cm for density of 1.2 Wb/m^2 in iron to be 5, the corresponding iron loss 10 be 2 ivátt/kg 50 Hz, and the density of iron as 7.8 kram/cm^3.

Solution. = 4.44x50x500*12xA

This is thenetcross-sectional area. However, the grossarea would be about for the insulation between laminations.

K = N/A', -40/500=4/50

∴NL, secondary voltage - KEI

750

80.0001=955

Total forgiven = 950+965=945.5

Max. value ofmagnetising current drawn by primary =845.5/500 = 1.691 A

Assuming this culTent to be sinusoidal, its is = I -691/ 1.196 A

Volumeofiron = length xarea=

Density 7.8 gram/em^3 Mass ofiron = 7.8/1(m-263.25kg

Total iron loss

Iron loss component Of no-load primary current is Iw =526.5/3000— O. 176 A

1.196^2 +0.1762 —0.208 A

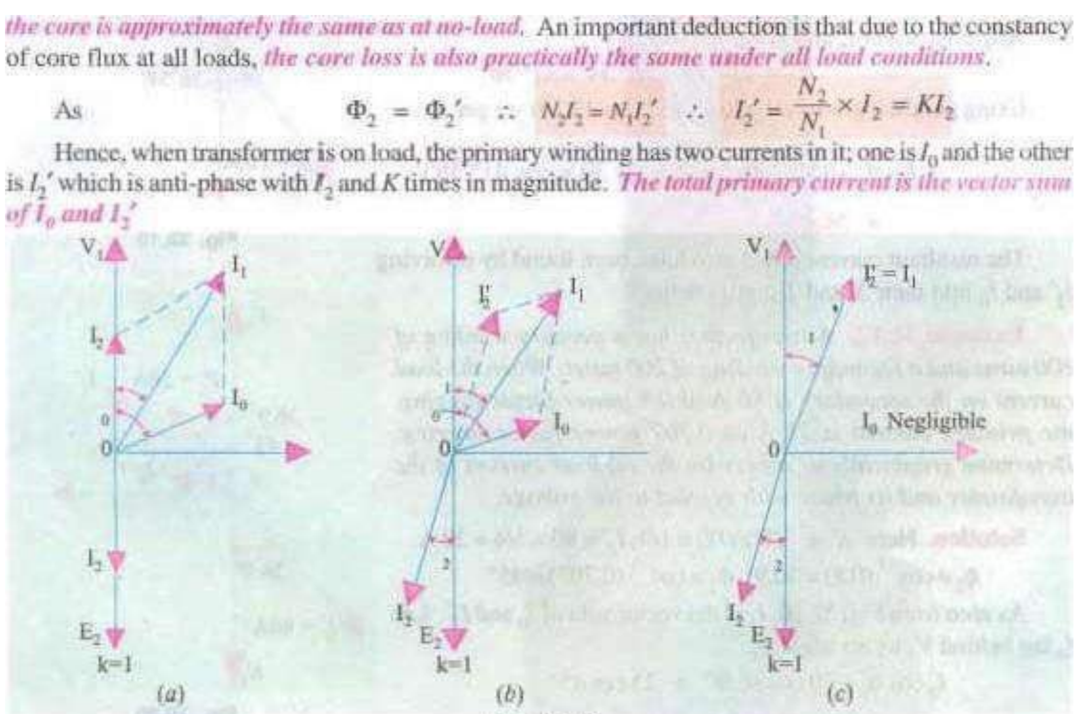
(d) powerfactor.cos,bo - 174=0.176/1208-0.1457

Transformer on Load 10

magnitude and phase of 12 with respect 'o V: is determined by the teristics ofthe Current', is in With V, ifload is non-inductive,it lags if load is inductive and it ieads ifload is cal•live.

The secondary current sets up its own and hence its own flux which is in opposition to the main primary flux which is due to The secondary ampere-turns N, 12 are known as demagnerising amp-turns. The opposing secondãry flux weakens the primary momentarily, hence primary back e.m_f- E, tendsto reduced. For a moment V, gains the uprrr hand over E, and hence causes more current to now in primary.

Let the additional primary current be V. It is known as load Ofprimary current. current is antiphase With V, The additional primary man-f. NI 12 sets up its own flux which is in opposition to CÞ, (but is in the same direction 4') and is equal to it in magnitude. Hence. the two cancel each other out. So. we find that the magnetic effects of secondary current 12 are immediately neutralized by the additional primary current l:' which is brought into existence exactly at the same instant as 12. The wholeprœess is illustrated in Fig. 32.17. Fig. 32.17


the core is approximately the same as at no-load. An important deduction is that due to the constancy of core flux at all loads, *the core loss is also practically the same under all load conditions.*

As $\Phi_2 = \Phi_2'$ $\therefore$ $N_2I_2 = N_1I_2'$ $\therefore$ $I_2' = \frac{N_2}{N_1} \times I_2 = KI_2$

Hence, when transformer is on load, the primary winding has two currents in it; one is I_0 and the other is I_2' which is anti-phase with I_2 and K times in magnitude. *The total primary current is the vector sum of I_0 and I_2'*

Fig. 32.18

Fig.

Hence, whatever the load conditions. the netflux passing through is approximately of core flux at all loads, the core loss

In Fig. 32.18 are shown the vector diagrams fora loaJtransformer when load is non-inductive and When it is.inductive(a similar diagram could be drawn for capacitive load). Voltage transformation ratio ofunity is assumed so that primary vectorsare equal to the secondary vectors. With reference to fig. 32.18 12 is secondary current in phase with E2 (strictly speaking it shouldbe 1/2). It causes primary current which is anti-phase With it and equal to it in magnitude (R = l). Total primary current I_1 is the vector sum Of and 12' and lags behind VI by an angle ϕ_1.

In Fig. 32.18 (b) vectors are drawn for an inductive load. Here 12 lags Ea (actually by ϕ_2. Current 12' is again antiphase with Gand equal to itin magnitude. As before, IOS the vectorsgmof/2' and 10 and lags behind V, by

It will be observed that isslightly greater than ϕ_2. But if we neglect I_0 a But if we neglect 10 as compared t042' Moreover.under under this assumption as in Fig. 32.18 then

$$N_2 I_1 = N_1 I_2 \quad \therefore \quad \frac{I_2'}{I_2} = \frac{I_1}{I_2} = \frac{N_2}{N_1} = K$$

NIG =

It shows that under full-load conditions, the ratio of primary and secondary currents is constanL This important relationship is made the basis of current transformer—a transformer which is used with alow-range ammeter for measuring currents incircuits Where the direct connection Of the ammeter is impracticable.

Fig. 3220

$E_1 = V_1 - I_1 R_1$ ...vector difference

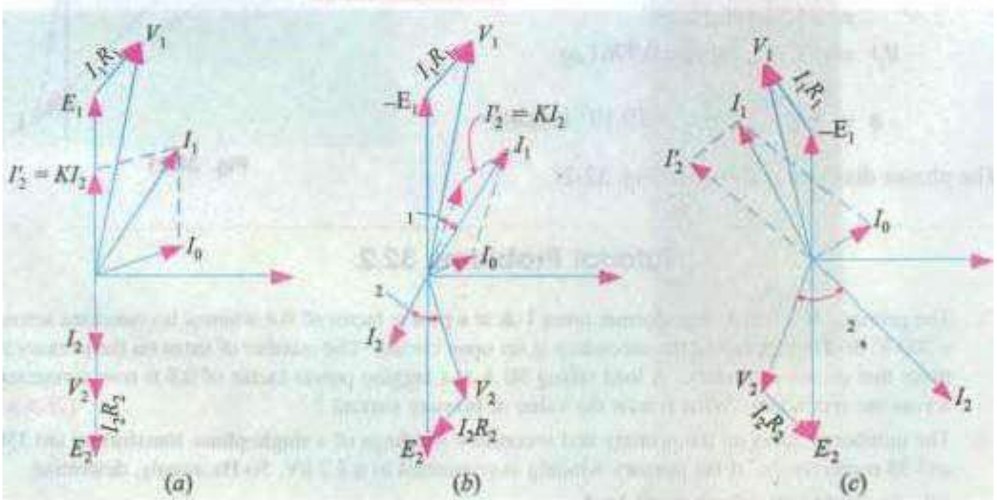


Fig. 32.22

The vector diagrams for non-inductive, inductive and capacitive loads are shown in Fig. 32.22 (a), (b)and (c) respectively.

Equivalent Resistance

In Fig. 32.23 a transformer is shown whose primary and secondary windings have resistances of RI and R: respectively. The resistances have been shown external to the windings.

It would now be shown that the resistances ofthe two windings can be transferredto any one ofthe two windings. The advantage of concentrating both the resistances in one winding is that it makes calculations very simple and easy because one has then to work in one winding only. It will be proved that a resistance of R: in secondary is equivalent to R2iK^2 in primary. The value Will be denoted by the equivalent Fig. 32.23 secondary resistance as referred to primary

The copper loss in secondary is 12^2 Ra, This loss is supplied by primary which takes a current ofll. Hence if R:' is the equivalent resistance in primary which would have caused the same loss as in secondary, then

Now. if we neglect 10, then 12/11 I/N , Hence, R/ R2/F

Similarly, equivalent primary resistance as refetïcd to secondary is R = I

In Fig. 32.24, secondary resistance has been transferred to primary side leaving secondary Circuit resistanceless. The resistance RI 4 R:' = RI + RI/K^2 is known as the equivalent or effective resistnnez Of

of the transformer as referred to primary and may be designated as R_{01}.

$$\therefore \quad R_{01} = R_1 + R_2' = R_1 + R_2 / K^2$$

Similarly, the *equivalent resistance of the transformer as referred to secondary is*

$$R_{02} = R_2 + R_1' = R_2 + K^2 R_1.$$

This fact is shown in Fig. 32.25 where all the resistances of the transformer has been concentrated in the secondary winding.

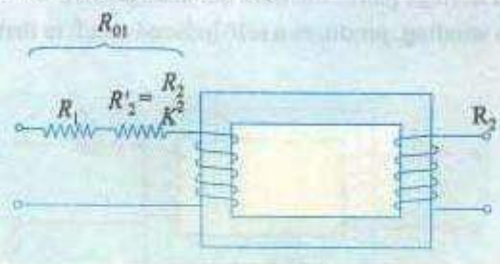


Fig. 32.24

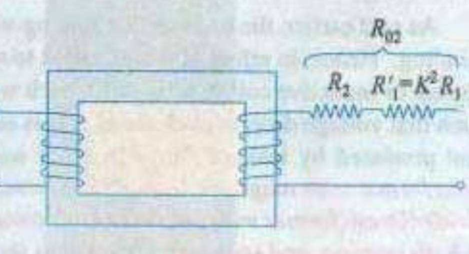


Fig. 32.25

the referred to primary and may be designated as

It is to be noted that

L a resistance of RI in primary is equivalent to FRI in secondary. Hence, it is called equivalent resistance as referred to secondary i. e. R I .

2. a resistance of R2 in secondary is equivalent to R2./K^2 in primary. Hence. it is called the equivalent secondary resistance as referred to primary e.

J. Total or effective resistance of the transformer as referred to primary is

Rot = primary resistance + equivalent secondary resistance as referred to primary

$$= R_1 + R_2' = R_1 + R_2/K^2$$

4. Similarly. total transformer resistance as referred to secondary is,

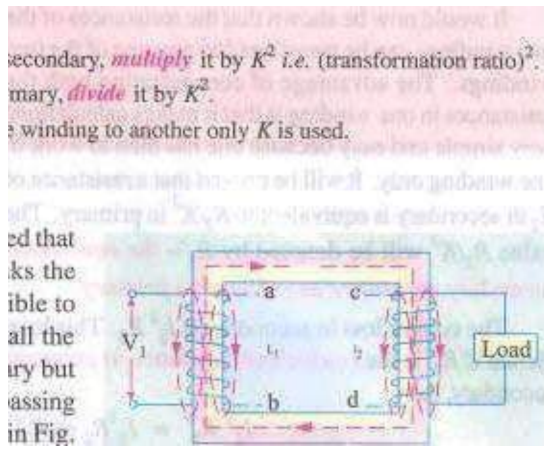
$$R_{02} = \text{secondary resistance + equivalent primary resistance as referred to secondary}$$
$$= R_2 + R_1' = R_2 + K^2 R_1$$

* Actually $I_2 \# 2/I_2' = 1/K$ and not $I_2 \# 2/I_1$. However, if I_0 is neglected, then $I_2' = I_1$.

Note : (tis important to remember that

(a) When shifting any primary resistance to the secondary, *nut/iply it by (b When shifting secondary resistance to the primary, divide it hy R.

(CJ however. when shifting any voltage from one.inding toanother



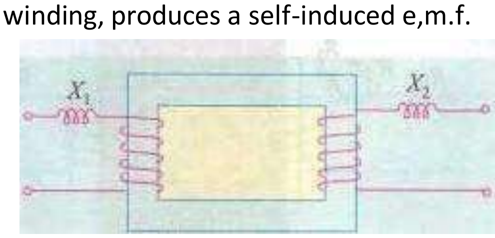
Magnetic Leakage

In the preceding discussion. it has been assumed that all the flux linked with primary winding also links the secondary winding. But, in practice, it is impossible to realize this condition. It is found, however, that all the flux linked with primary does not linkthe secondary but partofiti.e_OL completes itsmagnetic circuit bypassing through air than around the core, as shown in Fig. 32.26. This leakage flux is produced when the m.m.f. due to primary ampere-turns existing betueen points,a and Fig. 32.26 b, acts along the leakage paths. Hençe, this flux is known asprimary leakagejh'.r and is proportional to the primary ampere-turns alone because the secondary turns do not link the magnetic circuit Of The flux is in time phase with ll . It induces an e.m.f. in primary but notin secondary.

Similarly. secondary ampere-turns (or m.m.f.) acting acrosspoints c and d set up leakage flux Φ_{L} which is linked with secondary winding alone (and not With primary turns). This flux in time phase with 12 and produces a self-induced e.m.f. in secondary (but not in primary).

At no load and light loads. theprimary and secondary ampere-turns are small, hence leakage fluxes are negligible. But When load is increased. both primary and secondary windings carry huge currents. Hence. large m.m.f.s are Set up which, while acting on leakage paths, increase the leakage flux.

As said earlier, the leakage flux linking with each winding, produces a self-induced e,m.f. in that winding. Hence, in effect, it is equivalent to a small choker or inductive coil in series with each winding such that voltage drop in each series coil is equal to that produced by leakage flux. In other words, a transformer With



magnetic leakage is equivalent to an ideal transformer With inductive coils connecred in both primary and secondary circuils as shown in

Fig. 32.27 such that the internale.rn.f. in each inductive Fig. 32.27 coil is equal to that due to the corresponding leakage flux in the actual transformer.

$X_1 = e_{L1}/I_1$ and $X_2 = e_{L2}/I_2$

The terms X, and Xïare known as primary andsecondary leakage reactance-s respectively.

Following few points should be kept in mind :

1.The leakage flux links one or the other winding but no' both. hence it in nosvåy contributes to the transfer ofenergy from the primary to the secondary winding.

2. The primaryvoltage VI will havetosupply reactive drop/1X, in addition to Similarly E2 will have to supply 12 R2 and $I_2 X_2$.

3. In an actual transformer, the primary and secondary windings are not placed on separate legs Or limbs as Shown in Fig. 32,27 because due to their being Widely separated, large primary and secondary leakage fluxes would result. These leakage fluxes are minimised by sectionalizing and interleaving the primary and secondary windings as in Fig. 32.6 or Fig. 32.8.

Transformer with Resistance and Leakage Reactance

In Fig. 32.28 the primary and secondary windings Of a transformer With reactances taken out Of the windings are shown. The primary impedance is given by $Z_1 = \sqrt{(R_1^2 + X_1^2)}$

z, z:

Similarly, secondary impedance is given by

The resistance $Z_2 = \sqrt{(R_2^2 + X_2^2)}$ and leakage reactance Of each winding is responsible for some voltage drop in each winding. In primary, the leakage reactance drop is IlXi (usually I or 2% of VI). Fig. 32.28

Hence

Similarly, there are 12R2 and '2X2 drops in secondary which combine with V: to give El.

The vector diagram for such a transformer for different kinds ofloads is shown in fig. 32.29. In these diagrams, vectors for resistive drops are drawn parallel to current vectors whereas reactive drops are perpendicular to the current vectors. The angle Vi and gives the power factor angle of the transformer.

It may be noted that leakage reactances can also transferred from one winding to the other in the same way as resistance.

$$X_2' = X_2/K^2 \quad \text{and} \quad X_1' = K^2 X_1$$
$$X_{01} = X_1 + X_2' = X_1 + X_2/K^2 \text{ and } X_{02} = X_2 + X_1' = X_2 + K^2 X_1$$

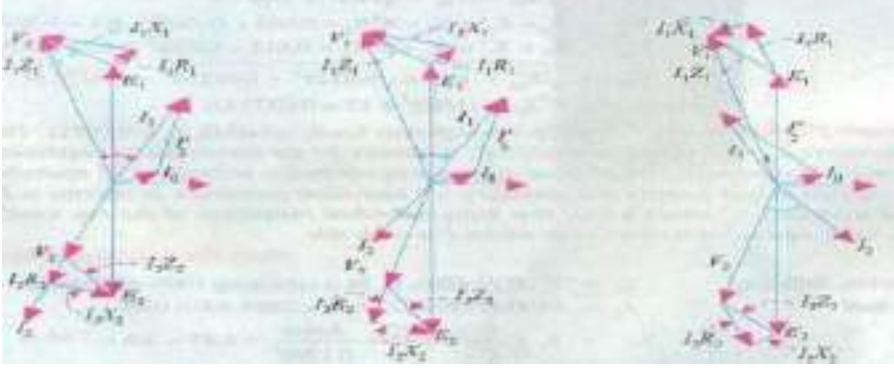


Fig. 32.29

Itis obvious that total impedanceofthe transformer as referred to primary is givenby

32.30 (a) $$Z_{01} = \sqrt{(R_{01}^2 + X_{01}^2)}$$ ...Fig.

Fig. 32.30 (b) $$Z_{02} = \sqrt{(R_{02}^2 + X_{02}^2)}$$ ...Fig.

Example 32.15. A 30 kVA, 2400/120-V, 50-Hz transformer has a high voltage winding transformer has a high voltagewinding resistance Of0.1 and a leakage reactance of 0.22Q. The IOW voltage winding resistance is 0.035 Q and the leakage reac'ance is Q Find the equivalent winding resistance, reacrance and Impedance referred to the O) high voltage side and (it) the low-voltage side.

(Electrical Machines-I, Bangalore Univ. 1987)

Solution. $K = 120/2400 = 1/20; R_1 = 0.1\ \Omega, X_1 = 0.22\ \Omega$

$R_2 = 0.035\ \Omega$ and $X_2 = 0.012\ \Omega$

(*i*) Here, high-voltage side is, obviously, the primary side. Hence, values as referred to primary side are

$$R_{01} = R_1 + R_2' = R_1 + R_2/K^2 = 0.1 + 0.035/(1/20)^2 = 14.1\ \Omega$$
$$X_{01} = X_1 + X_2' = X_1 + X_2/K^2 = 0.22 + 0.12/(1/20)^2 = 5.02\ \Omega$$
$$Z_{01} = \sqrt{R_{01}^2 + X_{01}^2} = \sqrt{14.1^2 + 5.02^2} = 15\ \Omega$$

(*ii*)
$$R_{02} = R_2 + R_1' = R_2 + K^2R_1 = 0.035 + (1/20)^2 \times 0.1 = 0.03525\ \Omega$$
$$X_{02} = X_2 + X_1' = X_2 + K^2X_1 = 0.012 + (1/20)^2 \times 0.22 = 0.01255\ \Omega$$
$$Z_{02} = \sqrt{R_{02}^2 + X_{02}^2} = \sqrt{0.0325^2 + 0.01255^2} = 0.0374\ \Omega$$
$$(\text{or } Z_{02} = K^2Z_{01} = (1/20)^2 \times 15 = 0.0375\ \Omega)$$

Example 32.16. *A 50-kVA, 4,400/220-V transformer has $R_1 = 3.45\ \Omega$, $R_2 = 0.009\ \Omega$. The values of reactances are $X_1 = 5.2\ \Omega$ and $X_2 = 0.015\ \Omega$. Calculate for the transformer (i) equivalent resistance as referred to primary (ii)* equivalentresistanceas referred to seéondary(Y) equivalentreac-

equivalentimpedanceas referredto bothpri-

Example 32.16. A Of reactances are X, = 5.2 lance as referred to primary tance as referred ta both primary and secondary

'nary and secondary total Cu loss, first using individual tvsistances Of the two windings and secondly, using equivalent resistances as referred to each side.

Solution. Full-load

(Elect. Engg.-I, Nagpur Univ. 1993)

Full-load $I_1 = 50,000/4,400 = 11.36$ A (assuming 100% efficiency)

$I_2 = 50,000/2220 = 227$ A; $K = 220/4,400 = 1/20$

(*i*) $R_{01} = R_1 + \dfrac{R_2}{K^2} = 3.45 + \dfrac{0.009}{(1/20)^2} = 3.45 + 3.6 = 7.05\ \Omega$

(*ii*) $R_{02} = R_2 + K^2R_1 = 0.009 + (1/20)^2 \times 3.45 = 0.009 + 0.0086 = 0.0176\ \Omega$

Also, $R_{02} = K^2 R_{01} = (1/20)^2 \times 7.05 = 0.0176\ \Omega$ (check)

AlsoCu 10»

Total Approximate Voltage Drop in a Transformer

When the transformer is On no-toad then VI is approximately equal to El. Hence E2 = KEI = WI. Also. E2 = 0 V: where 0V2 is secondary terminal voltage on noload, hence no-load secondary terminal voltage is WI. The secondary voltage on load is V2.

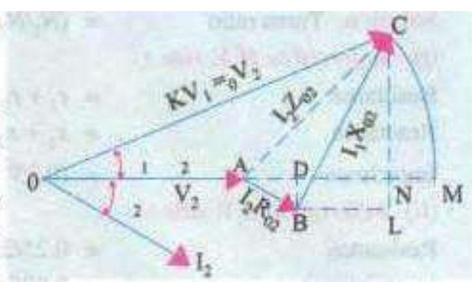


Fig. 32.35

utting OA produced at M. The total voltage drop I_2

The difference tween the two is 12 as shown in Fig. 32.35. The approximate voltage drop of the transformeras referredto secondary is found thus :

W'th O as the centre and radius OCdraw an arc cutting OAproduced at M. The total voltage drop 7-02 —AC AM which is approximately equal to AN. From B draw BD perpendicular on OA produced.

Draw CN to 0M and draw BL parallel to 0M.

Approximate voltage drop

= AN=AD+DN

+ 12Xcnsin $= I_2R_{02}\cos\phi$ φwhere $\phi_1 = \phi_2 = \phi$ (approx).

This is the value of approximate voltage drop for a laeging power factor,

The different figures for unity and leading power factors are shown in Fig. 32.36 (a) and (b) reqxctively.

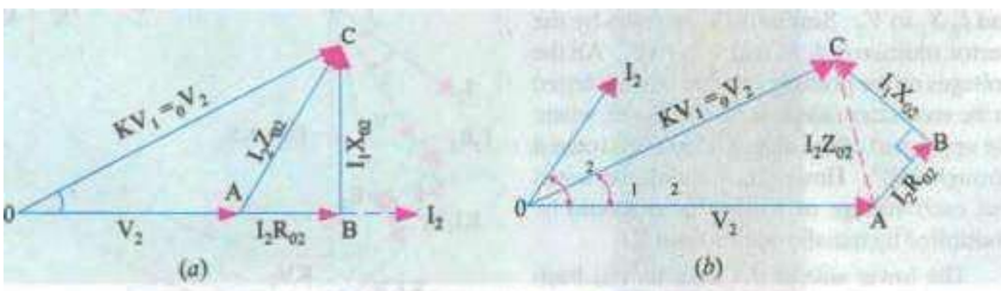


Fig. 32.36

The approximate voltage drop for leading rx»wer factor becomes

$(I_2R_{02}\cos\phi \pm I_2X_{02}\sin\phi)$

In general, approximate voltage drop is

ge drop is $(I_2R_{02}\cos\phi \pm I_2X_{02}\sin\phi)$

It may be noted that approximate voltage drop as referred to primary is

ate voltage drop as referred to primary is

$(I_1R_{01}\cos\phi \pm I_1X_{01}\sin\phi)$

% voltage drop in secondary

$$\text{is} = \frac{I_2R_{02}\cos\phi \pm I_2X_{02}\sin\phi}{{}_0V_2}\times 100$$

$$= \frac{100\times I_2R_{02}}{{}_0V_2}\cos\phi \pm \frac{100\,I_2X_{02}}{{}_0V_2}\sin\phi$$

= vrcos$*v.sino

$$v_r = \frac{100\,I_2R_{02}}{{}_0V_2} = \text{percentage resistive drop} = \frac{100\,I_1R_{01}}{V_1}$$

$$v_x = \frac{100\,I_2X_{02}}{{}_0V_2} = \text{percentage reactive drop} = \frac{100\,I_1X_{01}}{V_1}$$

Exact Voltage Drop

With reference to Fig. 3235, it is to be noted that exact voltage drop isAM and notAÑ, If we add the quantity NM to 'W, we will get theexact valueof the voltage drop.

Considering

Considering the right-angled triangle OCN, we get

$$NC^2 = OC^2 - ON^2 = (OC + ON)(OC - ON) = (OC + ON)(OM - ON) = 2\,OC \times NM$$

$$\therefore \quad NM = NC^2/2.OC \quad \text{Now, } NC = LC - LN = LC - BD$$

$$\therefore \quad NC = I_2 X_{02} \cos\phi - I_2 R_{02} \sin\phi \quad \therefore NM = \frac{(I_2 X_{02} \cos\phi - I_2 R_{02} \sin\phi)^2}{2_0 V_2}$$

$\therefore$ For a *lagging* power factor, exact voltage drop is

$$= AN + NM = (I_2 R_{02} \cos\phi + I_2 X_{02} \sin\phi) + \frac{(I_2 X_{02} \cos\phi - I_2 R_{02} \sin\phi)^2}{2_0 V_2}$$

For a *leading* power factor, the expression becomes

$$= (I_2 R_{02} \cos\phi - I_2 X_{02} \sin\phi) + \frac{(I_2 X_{02} \cos\phi + I_2 R_{02} \sin\phi)^2}{2_0 V_2}$$

In general, the voltage drop is

$$= (I_2 R_{02} \cos\phi \pm I_2 R_{02} \sin\phi) + \frac{(I_2 X_{02} \cos\phi \pm I_2 R_{02} \sin\phi)^2}{2_0 V_2}$$

Percentagedropis the

right-angled triangle OCN, we get

– 00

20v/ (12Ro, co. ± sin cos sin

2

= (v, COS ± v, sin O) + (1/200) (vx V, sinO)

The upper signs are to be used for a logging power factor and the lower ones for a leading power

Example 32.21. A 230/460-8/ transformer has a primary resistance of 0.2 and reactance Of 0.5 Q and rhe corresponding values for the secondary are 075 Q and 1.8£2 respectively. Find the secondary terminal voltage when supplying 10 A at 0.8 p.f. lagging.

(Electric. Machines-II, Bangalore Univ. 1991)

$$K = 460/230 = 2;\ R_{02} = R_2 + K^2 R_1 = 0.75 + 2^2 \times 0.2 = 1.55\ \Omega$$

$$X_{02} = X_2 + K^2 X_1 = 1.8 + 2^2 \times 0.5 = 3.8\ \Omega$$

$$\text{rop} = I_2 (R_{02} \cos\phi + X_{02} \sin\phi) = 10(1.55 \times 0.8 + 3.8 \times 0.6) = 35.2\text{V}$$

$$\text{age} = 460 - 35.2 = 424.8\ \text{V}$$

Solution.

Voltagedrop

∴Secondary terminal voltage =

Example 32.22. Calculate the regulation ofa transformer in which the percentage resistance drops is 1.0% and percentage reactance drop is 5.0% when the power factor is (a) 0.8 lagging (b) unity and (c) 0.8 leading. (Electrical Engineering, Bannras Hindu Univ. 1988)

Solution. We will use the approximate expression of Art 30.16.

(*a*) p.f. = cos ϕ = 0.8 lag $\quad \mu = v_r \cos\phi + v_x \sin\phi = 1 \times 0.8 + 5 \times 0.6 = 3.8\%$

(*b*) p.f. = cos ϕ = 1 $\quad \mu = 1 \times 1 + 5 \times 0 = 1\%$

(*c*) p.f. = cos ϕ = 0.8 lead $\quad \mu = 1 \times 0.8 - 5 \times 0.6 = -2.2\%$

Example 32.23. A transformer has a reaciance drop of5% and a resistancedmp of Find the lagging power factor at which the voltage regulation is maximum and the value of 'hi' regulation,

(Elect. FAEgg. Punjab Univ. 1991)

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$$\mu = v_r \cos\phi + v_x \sin\phi$$

where v_r is the percentage resistive drop and v_x is the percentagereactivedrop.

Differentiating the above equation, we get $\frac{d\mu}{d\phi} = -v_r$

For regulation to be maximum, $d\mu / d\phi = 0 \quad \therefore \quad -v$

or $\tan\phi = v_x/v_r = 5/2.5 = 2 \therefore \phi = \tan^{-1}(2) = 63.5°$ N

Solution. The percentage voltage regulation (p) is given by

sine. v coso

— vrsin vs cos = O

635^0 Now, cos = 0.45 and sin = 0.892

Maximum percentage regulation = (2.5 x 0.45) + (5 x 0.892) 5 $85

Maximum percentageregulation is 5.585 and occurs ata power factor of0Æ5 (lag).

Example 32.24. Calculate the percentage voltage drop for a transformer with a percentage resistance of 2.5% and.' reactance of 5% of rating 500 kVA when ir is delivering 400kVA at lagging, 'Elect. Machinery-I,

(%R) I cosO• Solution. % drop

Where If is the full-load cutTCnt and / the actual current.

In the present case,

Equivalent Circuit

The transformer shown diagrammatically in Fig. 32.37 (a) can be resolved into an equivalent circuit in which the resistance and leakage reactance of the transformer are imagined to be external to the winding whose only function then is to transform the voltage (Fig. 32.37 (b)). The no-load

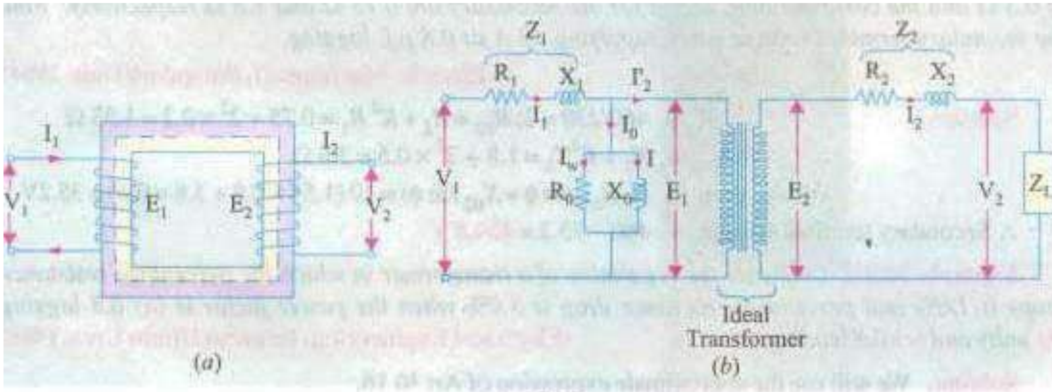


Fig. 32.37

current 10 is simulated by pure inductance Xo taking the magnetising component lg and a non-inductive resistance Ro taking the working component Is, connected in parallel across the primary circuit. The value of El is obtained by subtracting vector-idly 7-1 from VI. The value of Xo = El/loandof Ro = It is clear that El and E, are related to each other by expression

To maketransformer calculations simpler, it is preferable to transfer voltage, current and impedance

citherto the primary Or to the secondary. In thatcase. we would have to work in onewinding only which is more convenient.

The primary equi ofthe secondary induced voltageis E2' E2jK= El.

Similarly. primary equivalent Ofsecondary terminal or Output voltage is V,' V2/K Primary equivalentofthe secondary current is 12' = K12.

For transferring secondary mpedance to primary K^a is used.

$$R_2' = R_2/K^2,\ X_2' = X_2/K^2,\ Z_2' = Z_2/K^2$$

Thesamerelationship is Lised for shifting an external load impedance to the primal)'.

The secondary circuitis shown in Fig. 32.38(O) and its equivalent primary values areshown infig.

32.38 (b).

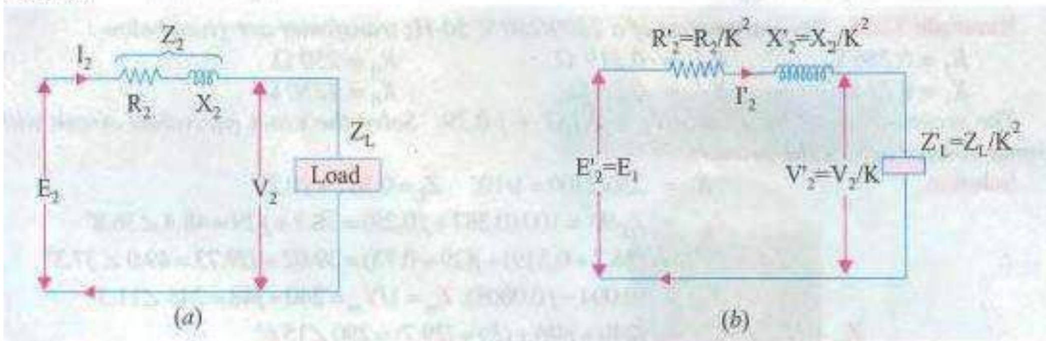


Fig. 32.38

The total equivalent circuit of the transformer is obtained by adding in the primar, impedance as shown

The total equivalent circuit ofthetransformeris obtainedby adding in the in Fig. 32.39. This is kncrwn as the exact equivalent circuit but it presents a somewhat harder circuit problem to solve. A simplification can be made by transferring the exciting circuit across the terminals as in Fig. 32.40 or in Fig. 32.41 (*a*). It should be noted that in this case $X_0 = V_1/I_\mu$.

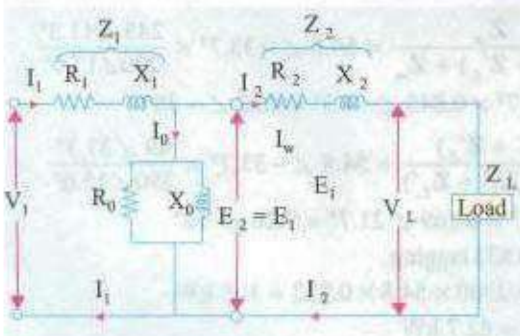


Fig. 32.39

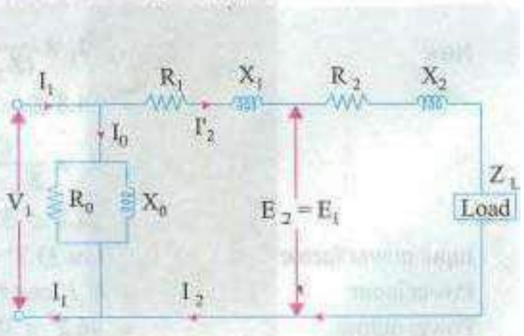


Fig. 32.40

Further simplification may be achieved by omitting I_0 altogether as shown in Fig. 32.41(*b*).

From Fig. 32.39 it is found that total impedance between the input terminal is

$$\mathbf{Z} = \mathbf{Z}_1 + \mathbf{Z}_m \,||\, (\mathbf{Z}_2' + \mathbf{Z}_L') = \left(\mathbf{Z}_1 + \frac{\mathbf{Z}_m\,(\mathbf{Z}_2' + \mathbf{Z}_L')}{\mathbf{Z}_m + (\mathbf{Z}_2' + \mathbf{Z}_L')}\right)$$

where $\mathbf{Z}_2' = R_2' + jX_2'$ and $\mathbf{Z}_m$ = impedance of the exciting circuit.

can bemade by transferring the exciting circuit aero:'S the tenninalsasin

This is so because there are two parallel circuits, one having an impedance ofZin andthe other having Z' and ZL' in series with each other.

$$\mathbf{V}_1 = \mathbf{I}_1\left[\mathbf{Z}_1 + \frac{\mathbf{Z}_m\,(\mathbf{Z}_2' + \mathbf{Z}_L')}{\mathbf{Z}_m + (\mathbf{Z}_2' + \mathbf{Z}_L')}\right]$$

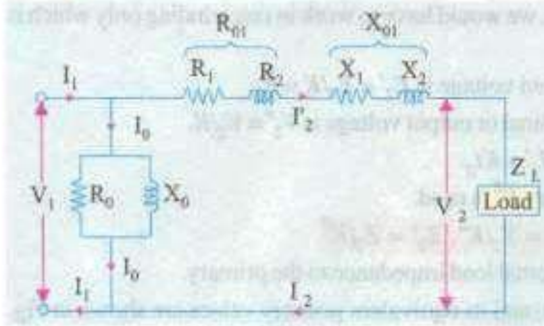


Fig. 32.41 (a)

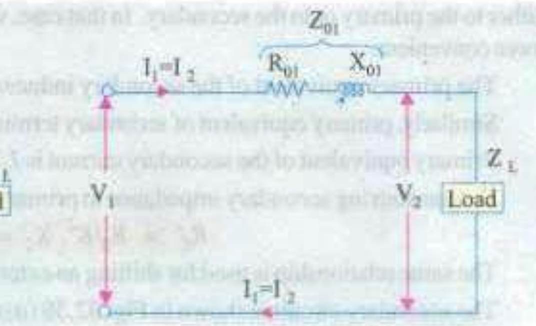


Fig. 32.41 (b)

Example 32.25. *The parametres of a 2300/230 V, 50-Hz transfomer are given below :*

$R_1 = 0.286\ \Omega$ $\quad R_2' = 0.319\ \Omega$ $\quad R_0 = 250\ \Omega$

$X_1 = 0.73\ \Omega$ $\quad X_2' = 0.73\ \Omega$ $\quad X_0 = 1250\ \Omega$

The secondary load impedance $\mathbf{Z}_L = 0.387 + j\,0.29$. *Solve the exact equivalent circuit with normal voltage across the primary.*

Solution.

$$K = 230/2300 = 1/10; \quad Z_L = 0.387 + j0.29$$

$$\mathbf{Z}_L' = \mathbf{Z}_L/K^2 = 100\,(0.387 + j\,0.29) = 38.7 + j29 = 48.4\,\angle 36.8°$$

$$\therefore \quad \mathbf{Z}_2' + \mathbf{Z}_L' = (38.7 + 0.319) + j(29 + 0.73) = 39.02 + j29.73 = 49.0\,\angle 37.3°$$

$$Y_m = (0.004 - j\,0.0008);\ \mathbf{Z}_m = 1/\mathbf{Y}_m = 240 + j48 = 245\,\angle 11.3°$$

$$\mathbf{Z}_m + (\mathbf{Z}_2' + \mathbf{Z}_L') = (240 + j48) + (39 + j29.7) = 290\,\angle 15.6°$$

$$\therefore \quad \mathbf{I}_1 = \frac{\mathbf{V}_1}{\mathbf{Z}_1 + \dfrac{\mathbf{Z}_m(\mathbf{Z}_2' + \mathbf{Z}_L')}{\mathbf{Z}_m(\mathbf{Z}_2' + \mathbf{Z}_L')}} = \left[\frac{2300\,\angle 0°}{0.286 + j0.73 + 41.4\,\angle 33°}\right]$$

$$= \frac{2300\,\angle 0°}{42\angle 33.7°} = 54.8\,\angle{-33.7°}$$

Now
$$\mathbf{I}_2' = \mathbf{I}_1 \times \frac{\mathbf{Z}_m}{(\mathbf{Z}_2' + \mathbf{Z}'_L) + \mathbf{Z}_m} = 54.8\,\angle -33.7° \times \frac{245\,\angle 11.3°}{290\angle 15.6°}$$

$$= 54.8\,\angle -33.7° \times 0.845\,\angle -4.3° = 46.2\,\angle -38°$$

$$\mathbf{I}_0 = \mathbf{I}_1 \times \frac{(\mathbf{Z}_2' + \mathbf{Z}'_L)}{\mathbf{Z}_m + (\mathbf{Z}_2' + \mathbf{Z}_L')} = 54.8\,\angle -33.7° \times \frac{49\,\angle 37.3°}{290\,\angle 15.6°}$$

$$= 54.8\,\angle{-33.7°} \times 0.169\,\angle 21.7° = 9.26\,\angle{-12°}$$

Input power factor $= \cos 33.7° = 0.832$ lagging

Power input $= V_1 I_1 \cos\phi_1 = 2300 \times 54.8 \times 0.832 = 105$ kW

Power output $= 46.2^2 \times 38.7 = 82.7$ kW

Primary Cu loss $= 54.8^2 \times 0.286 = 860$ W

Secondary Cu loss $= 46.2^2 \times 0.319 = 680$ W; Core loss $= 9.26^2 \times 240 = 20.6$ kW

$$\eta = (82.7/105) \times 100 = 78.8\%;\ V_2' = I_2' Z_L' = 46.2 \times 48.4 = 2{,}240\text{V}$$

$$\therefore \quad \text{Regulation} = \frac{2300 - 2240}{2240} \times 100 = \mathbf{2.7\%}$$

Example 32.26. *A transformer has a primary winding with a voltage-rating of 600 V. Its*

32.26.

transformer has a primary winding With a voltage-ratingOf 600 V. secondary-voltage rating is 1080 V With an additional tap a! 720 V. An 8 kW resistive load is connected across IOSO-V output terminals. A purely inductive load of 10k VA is connected across the tapping point and common second00' terminal so as get 720 V. Calculate the primary current and its power-factor Correlate if with the existing secondary loads. Neglect losses and magnetizing current. (Nagpur University, Winter 1999)

Solution. Loads are onnected as shown in Fig. 32142.

<u>8000</u>

=7.41 at unity p,f.

1080

= 10000/720=13.89atzerolaggingp.f.

These arereflected on to the primary sides with appropriate raùos ofturms, with corresponding powerfactors. If the conesponding transformed currents are represented by theabovesymbols modified by dashed superscripts,

$$I'_{r2} = 7.41 \times 1080/600 = 13.34 \text{ A at unity p.f.}$$

$$I'_{L_2} = 13.89 \times 720/600 = 16.67 \text{ A at zero lag. p.f.}$$

Hence, $$I_{r2} = [I'^2_{r2} + I'^2_{L2}]^{0.5} = 21.35 \text{ A, at 0.625 lag p.f.}$$

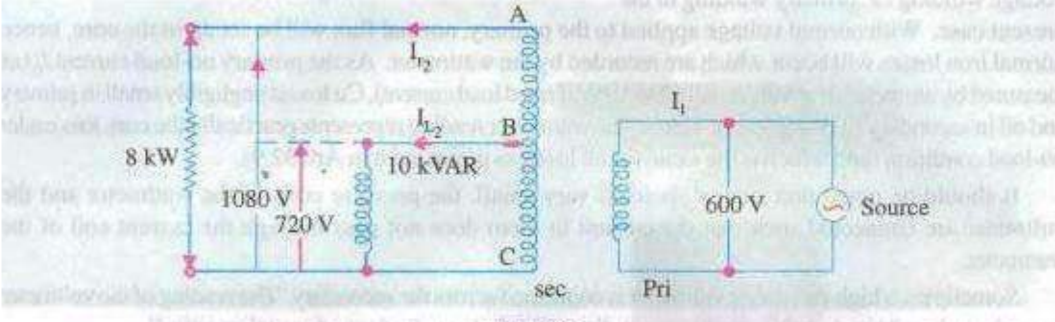


Fig. 32.42

Correlation : Since losses and magnetizing current are ignored, the calculations for primary current and its power-factor can also be made with data pertaining to the two Loads (in kW/kVAR), as supplied by the 600 V Source.

S = Load to be supplied : 8 kW at unityp.f, and I O kVAR lagging s = P+iQ=g-j 10kVA S =

Power—factor = $(8^2 + 10^2)^{0.5} = 12.8$ kVA

Primarycurrent = 12.8 x $\cos \phi = 8/12.8 = 0.625$ lag 1000/600=21.33 A

Transformer Tests

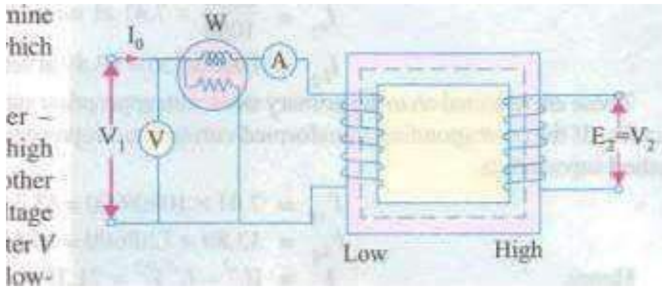
As shown in Ex 32.25. the periòrmanee ofa transfonnerean calculated on the basis of its equivalent circuit which contains (Fig. 32.41) four main parameters, the equivalent resistanceRot as referred to primary (or secondary R02), the equivalent leakage reactance as referred to primary (or secondary x02), the core-lossconductance Go (or resistance Roj und the magnetiSing susceptance BO (or reactance Xoj. These constants or parameters

can beeasily determined by two tests (i) open-circuit test and iil short. circuit test. These tests are very economical and convenient. because they furnish the required information without actually loading the transformer. fact, the testing of very large machinery consists of running two tests similar to the open and short-circuittests of atransformer.

Small transformer

Open-circuit or No-load Test

The purpose Of this test is to determine no-load loss orcore loss and no-load 10 which is helpful in finding xo and Ro,



One winding of the transformer — whichever is convenient but usually high vollage Winding — is left open and the Other is connected to its supply ofnormal voltage and frequency. A wattmeter W. voltmeter V and an ammeter A are connectedin the voltage winding i.e. primary winding in the Fig. 3243 present case. With normal voltage applied to the primary. normal fluxwill be set up in the core. hence normal iron losses Will which arc recorded by the wattmeter. As the primary no-load current 10 (as measured byammeter) is small (usually 2 to ofrated load cunænl), Cu loss is negligibly small in primary and nil in secondary fit being open). Hence. the wattmeter reading represents [Tactically thecole loss under no-load condition and which is same for all loads as pointed out in Art_ 32.9).

It should be noted that since is itself very small. the pressure coils Of the wattmeter and the voltmeter are connected such that the in them does not pass through the culTent coil Of the

Sometimes, connected acrossthe secondary. Thereading Ofthe voltmeter gives the induced e-m.f. in thesecondary winding._ Thishelps to find transformation ratio K.

The no-load vector diagramis shown in Fig. 32.16. If W is the wattmeter reading (in fig. 32,435.

W = VI locos \$0 :. cosOo= VWVl /o

= losin\$o. and

Or since the current is practically all-exciting current when a is on no-load (i.e. 10 and as the voltage drop in primary leakage impedance is hence the exciting admittance YO of the transformer is given by 10 V, Yo or Yo = WV,.

The exciting conductance Go is given by VI^2 Go or GO =

The exciting susceptance = (Y0^2 — G)

Example. 3227. In no-load rest of single-phase transformer, the following tesr data were obtained :

Primary voltage : 220 V : Secondary volt"ge : 110 V : primary current : 0.5 A ; Power input : 30 W.

Find thefollowing:

(i' The turns ratio (in the magtierisingcomponouofno-loadeurrent(iii' its working (orloss) component ii the iron loss.

Resistance ofthe primary winding = 0.6ohm.

Draw the no-load phasor diagram to scale. (Elect. Machine A.M.1-F„A99t))

Solution.

(iiij Gcos.-O.5x0.273—O.1365A

PrimaryCuloss = 0:52 x 0.6£0.15 w Iron loss = 30-0.15=29.85 w

Example 32.28.' 5 WA 200/1000 V50 single-phasefransfOrmeÈ gave results .

SC.

O) Calculate the parametersof the equivalent circuit rcferred 10 the L side.

(ii) Calculafe 'he output secondary voltage delivering 3 kW primary being 200 V, Find (he percènrakg regulation also.

(Nagpur University, November 19%)

Solution. (i) Shunt branch parameters from O.C. test (L.V. side) :

- 200/90 =4440hms, 200/444 = 0.45 amp = 1.1 1 amp, & = 200/1.11

All these are referred to LV. side.

(ii) Series.hranch parameters from S.C test (H.V side):

Since the S.C. test has been conducted from H.V. side, the parameters will refer to H.V. side. They should be converted to the parameters referred to L.V. side by transforming them suitably.

From S.C. Test readings,

These are referredto H,V' side.

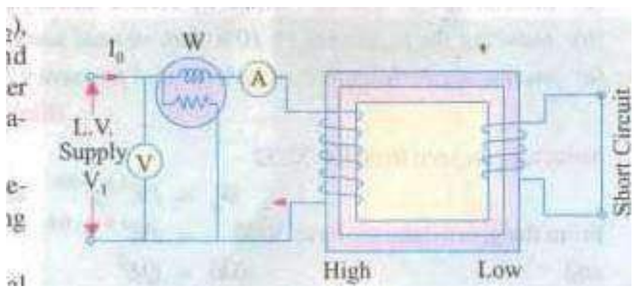
Equivalent circuit can be drawn with Round Xm calculated.tbove and r, and Xl.s above.

L V. Current at rated load= 5000/200 = 25 A

LN. lagging = 18.75 A

3222 Short-Circuit or Impedance Test

This is an economical method for determining the following : iii EAuivalent irnrrdance or G), leakage reactance (Xo, or Xo:) and total resistance (orR,z) ofthe transformer as referred to the winding in which the meaSuring irwtruments are placed.



(iii Cu loss at full load (andatanydesired load). 'This loss is used in calculating the efficiency ofthe transformer liii' Knowing Zo, or Zo:, the total voltage drop in the transformer as referred Fig. 32.45 to primary or secondary can be calculated and hence regulation Ofthe transformerdetermin

In this test. one winding, usually the low-voltage winding, issolidly short-circuited by a thickconductor (or through an ammeter Which may serve the additional Of indicating rated load curtrnt) as shown in FIE, 32.45.

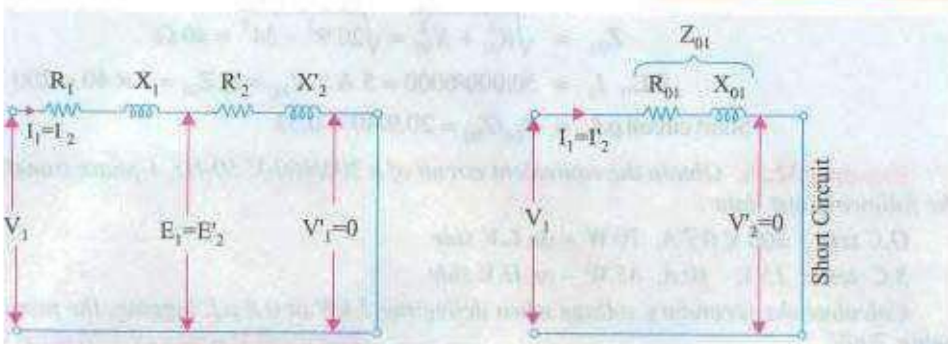


Fig. 32.46

A low voltage (usually 5 to arnormal primary voltage) at correcl frequency (though forCu losses it is not essential) is applied to the primary and is cautiously increased till full-loadcurrents are flowing both in primary and secondary indicated by the respecti ve ammeters).

Since, in this test. the applied voltage is a small percentage of the normal voltage, the mutual flux tÞ produced is also asmall percentage of its normal value (Art. 32.6), Hence. core losses are very small With the result that the wattmeter reading represent the full-load Cu loss or

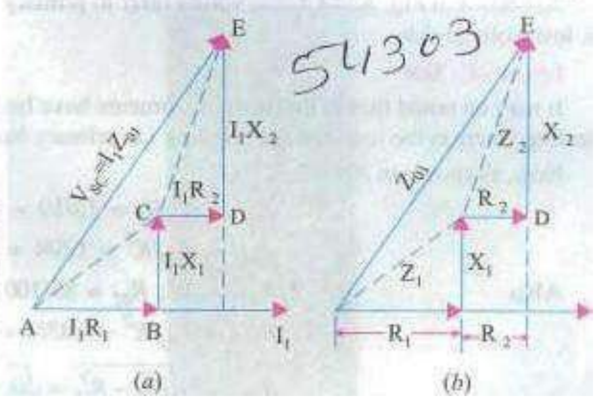
condition is shown in Fig. 32.46. If V_{sc} is the voltage required to circulate rated load currents, then Z_{01} V_{sc}/I_1

Also $\quad W = I_1^2 R_{01}$

$\therefore \quad R_{01} = W/I_1$

$\therefore \quad X_{01} = \sqrt{\left(Z_{01}^2 - R_{01}^2\right)}$

In Fig. 32.47 (a) the equivalent circuit vector diagram for the short-circuit test is shown. This diagram is the same as shown in Fig. 32.34 except that all the quantities are referred to the primary side. It is obvious that the entire voltage V_{SC} is consumed in the impedance drop of the two windings.



If R, can bemeasured, thenknowing ROI, wc can find R2' Ro − RI. The

1^2R loss forthe whole transformer i.e. both primary Cu loss and secondary Cu loss, The equivalent circuitof the tramsformer under shon-ciœuit condition is shown in fig. 32.46, If is the voltage required to =

Fig. 32.47 impedance triangle can then be divided into the appropriate equivalent triangles for primary and secondary as shown in Fig. 32.47(b).

32.23. Why Transformer Rating in kVA ?

As seen, Cu loss Of a transformer depends On current and iron loss on voltage. Hence. total transformer loss depends on volt-ampere (VA) and not on phase angle between voltage and cunent i.e. it is independent of load power factor. That is why rating of transformers is in kVA and not in kW, Example 32—35. The primary and secondary Windings of a 30 kVÄ 76000/230, V, I-phase transformer have resistance Of 0.016 ohm respectivelv. The mactance ofthe transformer referred to the primary is 34 Ohm.

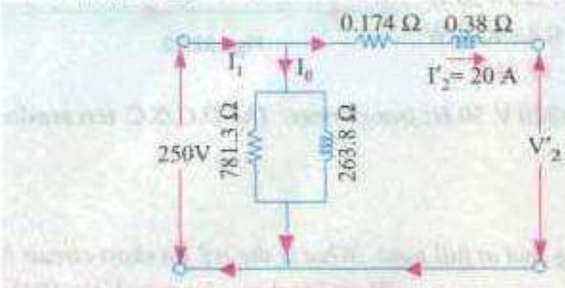


Fig. 32.53 (a)

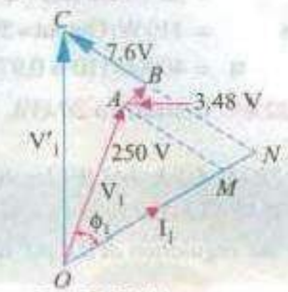


Fig. 32.53 (b)

$$\therefore \quad R_{02} = \frac{\text{short-circuit power}}{\text{F.L. secondary current}} = \frac{100}{12^2} = 0.694\ \Omega$$

$$Z_{02} = 20/12 = 1.667\ \Omega\ ;\ X_{02} = \sqrt{(1.667^2 - 0.694^2)} = 1.518\ \Omega$$

As R_0 and X_0 refer to primary, hence we will transfer these values to primary with the help Of transformation ratio.

$$K = 500/250 = 2 \quad \therefore \quad R_{01} = R_{02}/K^2 = 0.694/4 = 0.174\ \Omega$$

$$X_{01} = X_{02}/K^2 = 1.518/4 = 0.38\ \Omega\ ;\ Z_{01} = Z_{02}/K^2 = 1.667/4 = 0.417\ \Omega$$

The equivalent circuit is shown in Fig. 32.53 (*a*).

Efficiency

Total Cu loss $= I_2^2 R_{02} = 100 \times 0.694 = 69.4$ W ; Iron loss = 80 W

Total loss = 69.4 + 80 = 149.4 W $\quad \therefore \quad \eta = \dfrac{5000 \times 0.8 \times 100}{4000 + 149.4} =$ **96.42%**

The applied voltage V_1' is the vector sum of V_1 and $I_1 Z_{01}$ as shown in Fig. 32.53 (*b*).

$$I_1 = 20\ \text{A}\ ;\ I_1 R_{01} = 20 \times 0.174 = 3.84\ \text{V}\ ;\ I_1 X_{01} = 20 \times 0.38 = 7.6\ \text{V}$$

Neglecting the angle between V_1 and V_1', we have

$$V_1'^2 = OC^2 = ON^2 + NC^2 = (OM + MN)^2 + (NB + BC)^2$$

$$= (250 \times 0.8 + 3.48)^2 + (250 \times 0.6 + 7.6)^2$$

$$V_1'^2 = 203.5^2 + 157.6^2 \quad \therefore \quad V_1' = \mathbf{257.4\ V}$$

Example 32.48. *A 230/230 V, 3 kVA transformer gave the following results :*

O.C. Test : 230 V, 2 amp, 100 W

S.C. Test : 15 V, 13 amp, 120 W

Determine the regulation and efficiency at full load 0.80 p.f. lagging.

(Sambalpur 1998)

Solution. This is the case of a transformer with turns ratio as I : 1. Suchá transformer is mainly œquired for isolation.

Rated Current = 13 amp

230

Co-losses at rated load 120 watts.from S C. test

Core losses 100 Watts, from O.C. test

Atfull load. VAoutput - — 3000

At 0.8 lag p. Poweroutput

Required efficiency =

From S.C. test.

Approximate v01tage regulation

<u>13.51</u>

In terms of the voltage regulation = 230 x 100% -5.874%

Example 32.49. A 10 kVA, 500/250 V. single-phase transformer has its maximum efficiency of 94% When delivering Estimate its efficiency when delivering ifs full-load output p.f. of 0.8 lagging. (Nagpur UniveÑty, November 1998'

Solution. Rated output ag unity p.f. = 10000 W. Hence, of rated Output = 9,000 W

Input with efficiency =

Losses =

At maximum efficiency, variablecopper-loss = constant =Core 574/2 = 287 W loss = 574/2

At rated current, Let the copper-loss pc walls

At 90% load with unity p.f.. the copper-loss is expressed as 0.90^2 x Pi.

Hence, P, = 287/0.81=354W

(b) Output at full-load, 0.8 lag p.f. = 10,000 × 0.80 = 8000 W

At the corresponding load, Full Load copper-loss = 354 W

Hence. efficiency -8000/(8000 + 354 + 287) = 0.926-92.6<0

ofcalculation of voltage-magnitudes, approximav fOrrnuJa for voltage regulation can be used. For the present case of 0.8 lagging p.f.

VI' = V2+/[rcosQ+xsint$l

<u>230+43.5 (0.316 x</u>

= 230*43.5 (0.0634 + O. 18961—230+ 11 V, 241 volts.

It means that H.V. side terminal voltage must 2410 for keeping 230 V at the specified load.

(b) Approximate for voltage regulation is : V,' — =

With Laggingp.f., sign is retained. With leadingpoWer-factor. the—ve sign is applicable. For the voltage-regulation to be zero. only leading Bf. condition can prevail. sin O

= nix-0.0792/0316=0.25

= 140. cos 4=0.97 leading

Corresponding sino = Sin 0.243

H.V. terminal voltage required is23m V to maintain 230 V at since regulation condition is under discussion.

Example 32.51. A 5 kVA. 2200Q20 single-phase transformer has thefollowing parameters.

H. V. side r, = 3.4 Ohms. x, = 7.2 ohms

L V: side : r: = 0.028 ohms, 0.060 Ohms

Transformer is made to deliver rated current at 0.8 lagging p.f. to a load connected on the L V. side. If the load voltage is 220 Vt calculate the terminal voltage on H. V side

(Neglect the exciting current). (Rajiv Gandhi Tichnicnl University, Bhopal.Summer 2001)

Solution. Calculations may be done referring all the parameters the LV. side first. Finally, the voltage required onH.V. side can be obtained after transformation.

ref. to L V. side = 5000/220-22.73A

Total winding resistance ref. to L.V. side = $r_1' + r_2$ = $(220/2200)^2 \times 3.4 + 0.028$

Total winding-leakage-reactance ref. to L.V. side = $x_1' + x_2$

$$= (220/2200)^2 \times 7.2 + 0.060 = 0.132 \text{ ohm}$$

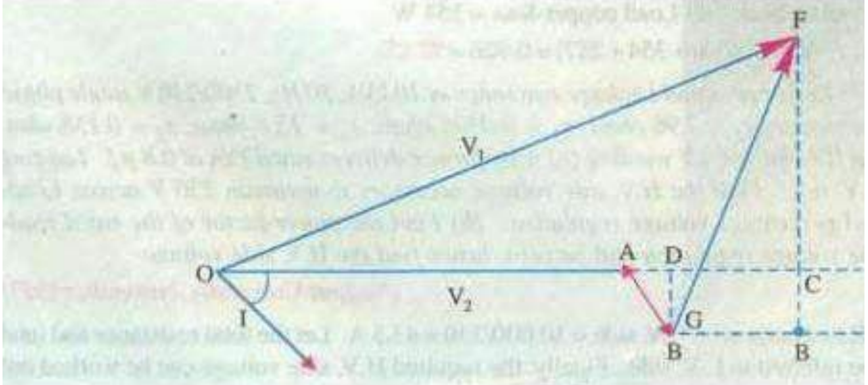


Fig. 32.53(c)

In the phasor diagram of Fig. 32.53 (*c*).

$$OA = V_2 = 220 \text{ volts}, I = 22.73 \text{ A at lagging phase angle of } 36.87°$$

$$AB = Ir,\ AD = Ir\cos\phi = 22.73 \times 0.062 \times 0.80 = 1.127 \text{ V}$$

$$DC = Ix\sin\phi = 22.73 \times 0.132 \times 0.60 = 1.80 \text{ V}$$

Rated

OC =220+ 1.127 + 1.80=222.93 volts

BD = 'r sin 4' = O. 85N

VF = xco.=2.40V

CF =240-0.85= 155 v

V,' = OF = $(222.93^2 + 1.55^2)^{0.50}$ = 222.935 volts

Required terminal voltage of H.V. side - VI - 222935 (2200/220) = 2229g5 volts

[Note. In approximate and fast calculations. CF is Often forcalculationof magnitude or The Concerned expression is: V, sin O, for lagging P.f.l

Example 32.52. A VA, 200/400 V. single-phase transformer takes O. 7amp and 65 W on Opencircuit. When the low-voltage Winding is short-circuited and / 5 V is applied '0 the high-voltage terminals. the current and power are TO A and 75 W respectively. Calculate thefun-load efficiency at unity factor and full•load regulation al 0.80 power-factor lagging.

(Nagpur UniversityApril 1999)

Solution. At a load of 4 kVA. the rated currents are : L.V side 4000/200 = 20amp

And H.V. side : 4000/400 = 10amp

From the test data. full-load copper-loss 75 AV

And Constant core-loss = 65 W

From S.C. test, Z = 15/10-1.Sohms

$R = 75/100 = 0.75$ ohm

$x = \sqrt{1.5^2 - 0.75^2} =$ = 130 | 1.52 — 0.752 | ohms

All these series-parameters are referred to the H.V. side, since the S.C. test has been conducted from H.V. side.

Full-load efficiency at unity p.f. = 40 / 65 75)

= 0.966-966%

Full load voltageregulation at 0.80 lagging p.f.

Thus, due to loading, H.M side voltage will drop by 16.14 volts (i.e. terminal voltage for the load will be 383.86 volts), when EV. side is energized by 200-V source.

Percentage Resistance, Reactance and Impedadce

These quantities are usually measured by the voltage drop át full-load current expressed as a percentage Of the normal voltage Of the winding on Which calculations are made.

'i' percentage resistance at full-load

x 100

% Cu loss at full-load

...Art-3216

(ii) Percentage reactance at full-load

(iii) Percentage impedanceat full-load

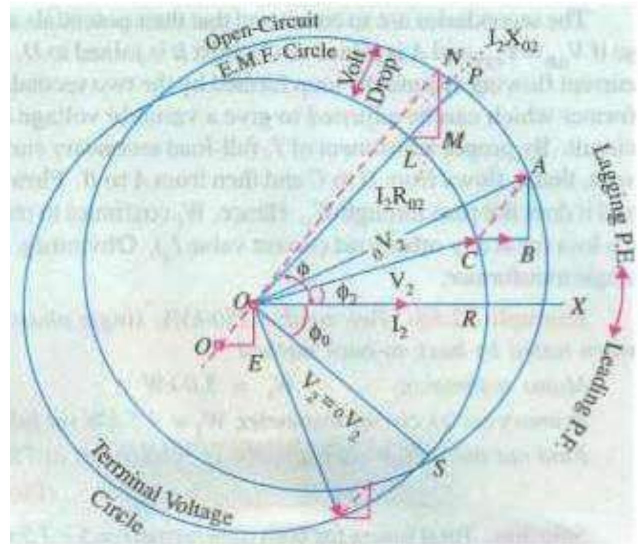


Fig. 32.54

It may be noted that percentage resistance, reactance and impedance have the same value whether referred to primary or secondary.

= 1.648+5.87=7.52 %

Example 32.57. A iràhsformer has coppeNossof and reactance-drop of3.5% when tested at full-load. Calculate its full-load regulation at (i) (ii) 0.8 pl. Lagging and (iii) 0.8 p'. Leading. (Bharathithasan Univ. April 1997'

Solution. The test-data at full-load gives following parameters :

p.a. resistance = 0.015, p.o. reactance —0.035

(t) Approximate Voltage — Regulation at unity p.f. full load

= 0.015cos040.035sin4'

=0.015 perunit=1.5%

(ii) Approximate Voltage — Regulationat 0.80 Lagging p.r.

(iii) Approximate Voltage Regulation

$= (0.015 \times 0.8) + (0.035 \times 0.6) = 0.033$ per unit $= 3.3\%$

lation at 0.8 leading p.f.

$= I_r \cos\phi - I_x \sin\phi$

$= (0.015 \times 0.8) - (0.035 \times 0.6) = -0.009$ per unit $= -0.9\%$ at 0.8

Kapp Regulation Diagram

It has been shown that secondary terminal voltage falls as the load on the transformer is increased whenp_f_ is lagging and it increases when the powerfactoris leading. In other words, secondary terminal voltage not only depends on the load but on power factoralso (Art. 32.16). For finding the voltage drop (Or rise) which is further used in determining the regulation Of the transformer, a graphical construction is employed which was proposed by late Dr. Kapp.

For drawing Kapp regulation diagram, it is necessary to know the equivalent resistance and reactance as referred to secondary i.e. R02 and x02_ If is the secondary load current, then secondary terminal voltage On load V2. is obtained by subtracting 12 and 12 x02 voltage drops vectorially from secondary no-load voltage 0V2.

NOW, o VI is constant, hence it represented by a circle Ofconstant radius OA as in 32.54. This circle is known as no-load or open-circuit e.m.f. circle. Fora given load. 012 represents the load current and is taken as the reference vector, CB represents 12 Rm and is parallel to 012,AB represents I_2 x02 and is drawn at right angles to CB. Vector OC obviously represents and is drawn at right anglestoCB. Vector OCobviously represents secondary terminal voltage Since 12 isconstant, the drop triangle ABC remains constant in sizz It is seen that end point C of V: lies on another circle whose centre is O'. This point (Y lies ata distance of 12 venically low the v»int O and a distance of l, Rm to its left as shown in Fig. 32.54.

Suppose it is required to find the voltage drop On full-load at a lagging cower factor ofcos then aradius OLP is drawn inclined at an angle of with OX. LM = 12 and is drawn horizontal MN = 12 x02 and is drawn perpendicular to 1M. Obviously, ON is noload voltage ov:. Now, ON OP ov2. Similarly. OLis V2. The voltage drop-op-OL=LP.

Hence. percentage regulation •down is — $\frac{OP - OL}{OP} \times 100 = \frac{LP}{OP} \times$ 100

It is seen that for finding voltage drop. triangle I-MN need not be drawn. but simply the radius OLD.

The diagram shows clearly how the secondary terminal voltage falls as the angle Of lag increases. Conversely, for a leading factor. the fall in terminal voltage decreases till foran angle of 00 leading, the fall becomes zero; hence V, = ov2. For angles greater than secondary terminal voltage V: greater than oV_2

The Kapp diagram is very helpful in determining the variation ofregulation with

the radii of thecircles. diagram has to be drawn on a very large scale, if sufficiently accurate results are

Sumpner Of Back-to-Back Test

This test provides data for finding regulation. efficie1EY and heating conditions and is employed only when two similar transformers are available. One is loaded on the other and both are connected to supply. The power taken from thesupply is thatnecessary for supplying the losses of both transfOrmers and the rEgligibly small in thecontrol circuit.

As shown in Fig. 3255. primaries of the two transformers are connected in parallel across the same a.c. supply. With switch S open, the wattmeter WI reads the core loss for thetwo t.ra1VSfornErs.

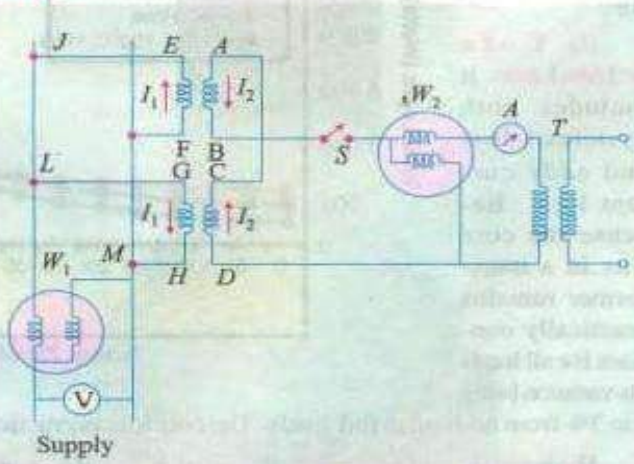


Fig. 32.55

powerfactorbutithas the disadvantage that since the lengths ofthe sides of the impedance triangle are small as compared to The secondaries are so connected that their potentials are in opposition to each other. This would so if VAB = vCD and A is joined to C whilst B is joined to D. In that case, there would be no secondary current flowing around the loop formed bythe two secondaries. Tis anauxiliary low-voltage transformer which can be adjusted to given variable voltage and hence current in thesecondary loop circuit. By proper adjustment of T, full-load secondary current can be made to flow as shown. It is seen, that 12 flows from D to C and then fromA to B. Flow of is confined to the loopFEJLGHMF and it does not pass through WI. Hence, W, continues toread the core loss and W2measures full-load Cu loss (or at any other load current value 12). Obviously, the power takenin is twice the lossesofa single transformer.

Example 32.58. Two similar 250-kVA, single-phase transformers when tested by back-to-back method:

MainsWI = 5.0 kW

Primary series circuit Wattmeter, W2 Z-5 kW fat full-load cu"rent), Find Old the individual transformer efficiencies at

Solution. Total losses for both transformers —5 + 7.5 = 12.5 kW EL. loss for each transformef = 12.5/2= 6.25 kW

$$\text{Copper-loss at 75\% load} = \left(\frac{3}{4}\right)^2 \times \frac{7.5}{2}\ \text{kW} = 2.11\ \text{kW}$$

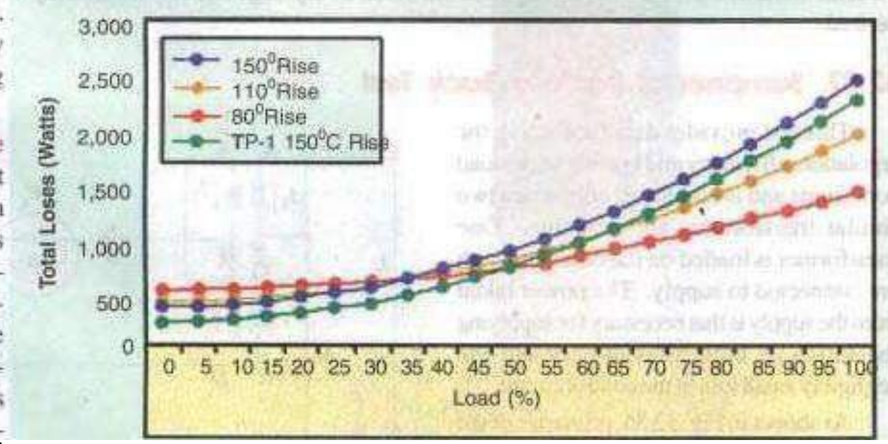
Output of each transformer at 75% F.L. and 0.8 p.f. = $(250 \times 0.75) \times 0.8 = 150$ kW

$$\eta = \frac{150}{150 + 2.5 + 2.11} = 97\%$$

32.28. Losses in a Transformer

In a static transformer, there are no friction or windage losses. Hence, the only losses occuring are :

(i) Core or Iron Loss : It includes both hysteresis loss and eddy current loss. Because the core flux in a transformer remains practically con-



.000

0

Stant for all loads   Typical 75kVA Transformer LossesNS, Load

(its   tx•ing

I to 3% from no-load to full-load). The core loss is practicallýthe sam at all loads.

Wh = $\eta B^{1.6}_{max} f$

These losses are minimized by using steelofhigh silicon content for the core and by using very thin laminations. Iron or core loss is found tbeO,Ç. (esc, The load measures core loss

(ii) Copper loss. This loss is   resistance

= 11^2R1 + Rot + Itis clear that Culossikproportionalto (current/ or kVA. In other words, Cu loss at half thelull-load is one-fourth ofthat at full-toad.

The value OfCUIOSS is found from the shon-circuit test (Art. 32.22).

32.29. Efficiency Of a Transformer

Asisthe case with other types ofelectrical machines, the efficiency of atransformer at a particular load and power factor isdefined as the outputdivided by the input—the two being measured in the same units (either watts or kilowatts)

$$\text{Efficiency} = \frac{\text{Output}}{\text{Input}}$$

But a transformer being ahighly efficient piece ofequipment, has very small loss, hence it is impracticål to to measure transformer, efficiency by measuring input and output. These quantities are nearlyofthe same sire, A better method is to determine the losses and then to calculate theefficiency from

$$\text{Efficiency} = \frac{\text{Output}}{\text{Output} + \text{losses}} = \frac{\text{Output}}{\text{Output} + \text{Cu loss} + \text{iron loss}}$$

$$\eta = \frac{\text{Input} - \text{Losses}}{\text{Input}} = 1 - \frac{\text{losses}}{\text{Input}}$$

Ou!pyt

It may be noted here that efficiency is based on power output in watts and not in volt-amperes, although losses are proportional to VA. Hence;at anyvolt-ampere load. theefficiency depends on power factor, beingmaximum at a power factorofunity.

no-load or open-circuit

Efficiency can be computed by determiningcore from testand Cu lossfrom the short-circuittest,

32.30. Condition for Maximum Efficiency

culoss

For nto be maximum.

The output current cormsrxmdingto maximum efficiency is 12 $= \sqrt{(W_i/R_{02})}$.

It is this value of theoutput current which will make the Cu loss equal to theironlosg. By design. it is possible to make the maximum efficiency occur at any desired load.

Iron loss
cu loss

Note. If we arc given iron loss and fullload Cu loss. then the load at Which two losses would equal (i.e. to maximum efficiency) is given by

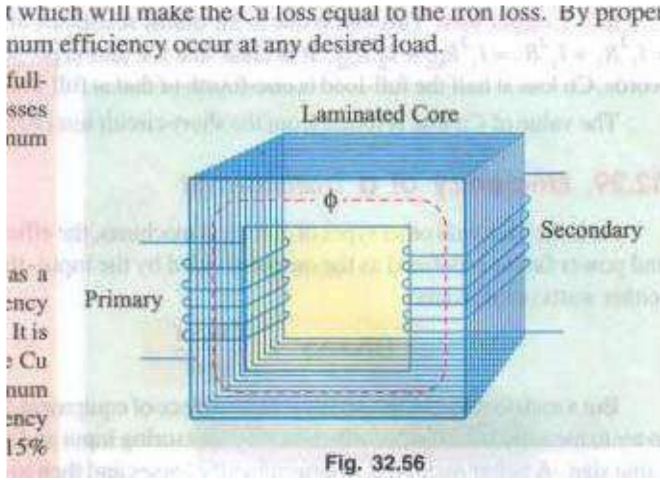


Fig. 32.56

-In Fig, 3256, Cu lóses are plotted tkm:ntagc Of power input and the cfficicngy curve as deduced from these is also shown. It is obvious that the point of intersection of the Cu and iron loss curves gives the point of maximum efficiency. It seen that the efficienćý is high and is practically constant from full-load to overload. (íi) The efficiency atany load is given by

where x

Example 323. In a 25-kVA, 2000/200 V. single-phase transformer. the iron andfull-load copper losses are 350 and 400 W respectively. Calculate the efficiency at unity power factor on

(i) full load (ii) halffull-load. (Elect. Engg. & Electronic. Bangalore Univ. 1990 and Similarexample in U.P. Technical University 2001)

Soluti'm. (i) Full-load Unity pl.

Total loss =

FL output at u.p.f. =25 x Half F.L. Unityp.r.

cu loss = 400 x (1/2/ = 100 W. Iron loss remains constant at 350 W. Total loss = 350 -450 w.

Half-load output at u.p.f. = 12.5 kW

Example 320. and p: be the iron andcopper losses ofa transformeronfull•load.findthe ratio OfP, and pz such that efficiency occurs at full-load.

Sec. B, Summer 1992)

Solution. If P: is theCu loss atfull-load, its valueat 75% offull.load is—p: At maximum efficiency, it equals the ironloss pa which remains constantthroughout. Hence, at maximum

efficiency.

p, 9P2/16 or P/Pz-9/t6.

Example 32.61. A 11000/230 V, 150-kVA. I-phase, transformer has core loss Of1.4 kW and EL Cu loss OfI kW Determine

(i) rhe WA loadfor max efficiency and value Of efficiency at unity p.f

(io rhe efficiency at halfEL 0.8p.f. leading (Basic Elect-Machine,NagpurUniv. 1993)

Solution. (i) Load WA correspondingto maximumefficiency is

Since Cu loss equals iron = F.L. kVA × loss at maximum efficiency, totul loss output = 160 x I = 160 kW

= 160/162.8 -0.982 or 9K2%

(ii) Cu lossat half full-load = 1.6k $(1/2)^2$ = 0.4 kW Total loss= 1.44 0.4 = 1.8 kW

HalfEL. output at 0.8 p.f.= (0/2) × 0.8 = 60 kW 150/2) FAIiciency =

Example 32.62. A 5-kVA. 2,30QQ30-u 50•Hz transformer was testedfor the iron losses 'With normal excitation and Cu losses atfull-loud and these werefound to be 40 W and 112 W 'espectively.

Calculate the efficiencies Of the transformer at 0.8 powerfactorfor thefollowing WA outputs :

1.25 2.5 3.75 so 6.25 7.5

Plot efficiency vs kVA output curve. **(Elect. Engg. -I, Bombay Univ. 1987)**

Solution. F.L. Cu loss $= 112$ W ; Iron loss $= 40$ W

(*i*) Cu loss at 1.25 kVA $= 112 \times (1.25/5)^2 = 7$ W

Total loss $= 40 + 7 = 47$ W Output $= 1.25 \times 0.8 = 1$ kW $= 1{,}000$ W

$\eta = 100 \times 1{,}000/1{,}047 =$ **95.51 %**

(*ii*) Cu loss at 2.5 kVA $= 112 \times (2.5/5)^2 = 28$ W

Total loss $= 40 + 28 = 68$ W

Output $= 2.5 \times 0.8 = 2$ kW

$\eta = 2{,}000 \times 100/2{,}068 =$ **96.71%**

(*iii*) Cu loss at 3.75 kVA

$= 112 \times (3.75/5)^2 = 63$ W

Total loss $= 40 + 63 = 103$ W

$\eta = 3{,}000 \times 100/3{,}103 =$ **96.68 %**

(*iv*) Cu loss at 5 kVA

$= 112$ W

Total loss $= 152$ W $= 0.152$ kW

Output $= 5 \times 0.8 = 4$ kW

$\eta = 4 \times 100/4.142 =$ **96.34 %**

(*v*) Cu loss at 6.25 kVA

$= 112 \times (6.25/5)^2 = 175$ W

Total loss $= 125$ W $= 0.125$ kW ; Output $= 6.25 \times 0.8 = 5$ kW

$\eta = 5 \times 100/5.215 =$ **95.88 %**

(*vi*) Cu loss at 7.5 kVA $= 112 \times (7.5/5)^2 = 252$ W

Total loss $= 292$ W $= 0.292$ kW ; Output $= 7.5 \times 0.8 = 6$ kW

$\eta = 6 \times 100/6.292 =$ **95.36 %**

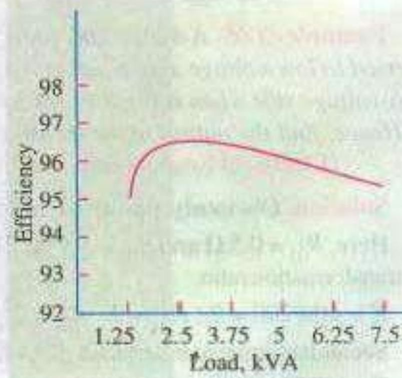


Fig. 32.57

The curve is shown in Fig. 32.57.

Example 32.63. 2ffl-kVA transformer has an efficiency Of ar full load. the mat. efficiency Occurs ar three quarters offull-load. calculate the efficiency at half Assume negligible magnetizing current and p.f. O,8 at all loads. (Elect. Punjab UniV. Jan. 1991)

Solution. As given. the transformerhas a EL efficiency of98 % pr.

El..output = 200x0.8= 160kW ; EL. input- 60/0.98 =163.265 kW

F.L.tosses = 163.265- 160=3.265kW ELoutput = 25XO.8=20kW- 20,OOOW

Full-load = 20,000 x 100/(20.000 464.4)= 9-7.' %

Example 32.6*. A 4-kVA. V, I-phase transfonnerhasequivalent n•sistance and œacrance referred ro low-voltage side equal to O, 512 1.5 Q respectively Find the terminal voltage On the high-voltage side When supplies 3/41/1full-load a/ powerfactor OfO.8, the supply voltage being 220 V. Hence, find the output transformer and its efficiency ifthe corelosses are 100 W.

(Electrical Engineering : Bombay Univ. 1985'

Solution. Obviously, primary is the side and the secondary. the high voltage side.

Here, ROI Q and = I S These can be transferred to the secondary side with the help Of the transformation ratio.

Secondary current when load is 3/4 the, full-load is (X4X 3/4)/400 7SA

Total drop as referred to transformer secondary is

Terminal voltageon high-voltage side under givenloadconditionis = 400-39=361 v culoss =
Iron loss 100 W

= 212SW output = (4 x 3/4) x 0.8 = 2.4 kW

Input = 2.400+212.5-2,612.5 w n-2.400x 100/2,612.5 = 91.87

Example 32.66. A 20•kVA, 440/220 V, 50 Hz transformer has iron loss W. The Cu isfound be When delivering halffull-load curt-em. Determine efficiency when

Assumingalagging factor

deliveringfull-load current at 0.8 lagging and /ii) the percent offull-load When the efficiency Will be marimum. (Electrutechnique•ll, MS. Univ., Baroda 1987)

Hence. efficiency would be maximum at90 % ofEL

Example 32.67. Consider a 4-kVA. 200/400 V Mingle-phase transformer supplying full-load current at 0.8 lagging powerfactor. The QC-IS. C. test results are as follows :

O.C.test : 200 V, O.SA. 70B' (I.V. side)

S.C. test 20M 10 A 60 V: Side'

Calculate efficiency, secondary voltage and current into primary at the above load.

Calculate the load at unity powerfactor corresponding to marimum efficiency iEleci. Machines Nagpur Univ. 1993È

Solution. Full-toad. = 4000/400 = 10 A

It means that S.C. test has been carried out with full Secondary flowing, Hence, 60 W represents full-load Cu loss of the transformer.

EL. losses -60 +

EL. n -3.2/3.33=0.960r96 %

Example 32.68. A 600 WA. I-phase transformer has an efficiency of92 % both atfu[l-loadand half-load at unity powerfactor. Determine its efficiency at 60 % offull-load at O.S powerfactor lag.

Sec. B,

x ICO

2

(x x WA) x cos + W, + .rWcu where x represents percentage Of full-load W, is iron loss and Wcu is full-load Cu loss,

At EL u.p.r. Here x =

W/+ -52.174kW

Atha1fFL UPE Her-ex=

*100;

= 85.9%

has an efficiency Of92 % atfull-load and also at half-load. Determine its efficiency When it operates at p.j: and 60 Of load. (Electric. Machines. Kerala Univ. 1987)

Solulion. The fact that efficiency is the same 92 % atbotb full-load and half-load Will help us to find the iron and copper losses.

At full-load

Output -600 kW ;

SinevCu loss becoloes one-fourthof its EL. value, hence x +y/4 - 26. l Solving for-rand y. we get 17.4kW ; y=

At 60 full-load cu 0.62 x34.s= 1253 kW ; Total loss- 17.4+ 12.53-29.93 kW

=360kW •02360/389.93=0.965or 96.5

Example 32.70. The maximum efficiency Ofa /OO-kVA, Single phase transformer is and occurs at offull load at 8 If rhe leakage impedance of transformer is find the voltage regulation al rated load Of O. 8 power factor lagging.

Elect. Machines-I, Nagpur Univ.1993)

Solution. Since maximum efficiency occurs at 80 percent Of full-load at 0.8 p.f..

64/0.98=65.3 kW

Culossat full-load = $0.65/0.8^2$ = I kW

Cu loss 1 00

x 100-lx

IOO

% age regn, = + 5 0.6)

Example 32.71. A 10WA, 5(YW440- single phase transformerhas eddyCurrent and hysteresis losses of I 0.5 and 0.6 per cent ofoutput onfull What "'ill be rhe percentage losses if the transformer is used on 50-11: system keeping the full-load current consJont ?

Assume unity powerfactor operation. Compare thefull load efficienciesfor rhe nvo cases.

(Elect. Machines, B, 1991

Solution. We know that El = 4.44 f When both excitation voltage and frequency are doubled, flux remains unchanged.

FL. output at upf 10kVA x I 10kW

FL Cu loss 1.5 x 10/100 = 0.1SkW : Eddy current loss

= 0.5 x = 0.05 kW : Hysteresis loss = 0.6 x

Now, full-load current is kept constant but voltage is increased from 5000 V to 10.000 V. Hence, output will be doubled to 20 kW. Dueto constant current, Cu loss would also remain constant.

New = 0.15 kW, % cu loss

Now, eddy curœntloss

New eddy current loss -0.05 IOO- 1 %

Now, = 0.06x (50/25) - 0.12 kW, %

at 0.8 p.f. a/ normal voltage

Solution.

Example 32.73. A single phake transformer is rated at IOO•kVA, 2300/230-V. 50 HZ.
The maximumflux density in the core is l. 2 Wb/m^2 and the net cross-sectional areo Ofthe core is O. 04"1^2.

Determine

(a) The number ofprimary and Secondary turns needed.

(b) [f the mean length Of the magnetic circuit is 2.5 m and the*elative permeability is 1200, determine the magnetising current. Neglect theßurrent drawn for the core loss.

(c) On short-circuit with full-load current flowing, the power input is 1200 W and an opencircuit With rated voltage, the power input was 400 W. Determine the efficiency of the transformer at 75 % offull-load With 0.8 p.f lag.

(d) Ifthe same transformer is connected to a supply thefrequency

(i.e.. 100 Whar is effect on ifs efficiency ? (Elect. Engg„ Bombay U"is. 19tB)

—9.21 A

Output— 100 x (3/4)

(d) When frequency is doubled. iron loss isincœased because

(i) hysteresis loss is doubled—

(ii) eddy current loss is quadrupled — We Hence. efficiency will decrease

what is rhe power-factor ar Which rhe regulation will•v' : '(i/ Zero; (ii) positive-maximum ? (b Ifits maximum efficiency occurs atfull-load (at unity p.f, what Will be efficiency under these conditions ?

Solution : Approximate perceptage regulation is given, in thi' case, by the relationship

I 5.4 sin O.

la) Regulation :

(i' If regulation is zero. negative sign must be applicable. This happens at leadings p.f.

Corresponding pf. = leading

= 18.44" leading

For maximum positive regulation, lagging p.f_ is a must. From phasordiagram, the can be obtained.

Corresponding tan 5.4/1.8

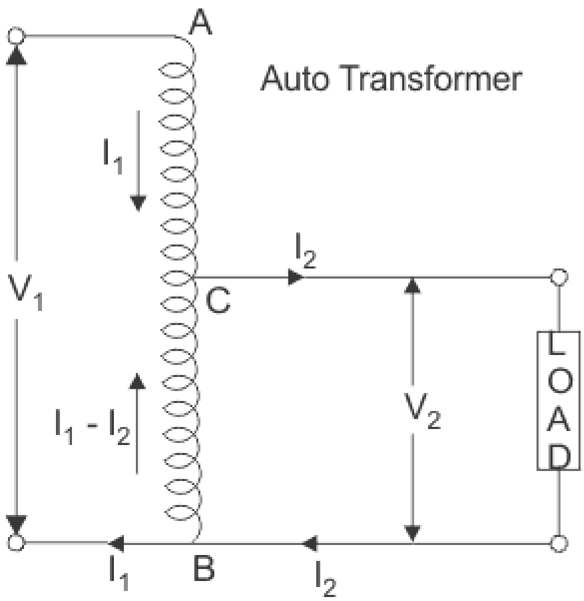
% Voltage regulation = 1.8cos4+5.4sin 0=5.7 % (b) Efficiency : Maximum efficiency occurs at such a Iron losses =

AUTO TRANSFORMER

An **autotransformer** is a kind of electrical transformer where primary and secondary shares same common single winding. So basically it's a one winding transformer.

Autotransformer Theory

In an auto transformer, one single winding is used as primary winding as well as secondary winding. But in two windings transformer two different windings are used for primary and secondary purpose. A circuit diagram of auto transformer is shown below.



The winding AB of total turns N_1 is considered as primary winding. This winding is tapped from point 'C' and the portion BC is considered as secondary. Let's assume the number of turns in between points 'B' and 'C' is N_2.

If V_1 voltage is applied across the winding i.e. in between 'A' and 'C'.

$So\ voltage\ per\ turn\ in\ this\ winding\ is\ \frac{V_1}{N_1}$

Hence, the voltage across the portion BC of the winding, will be,

$\frac{V_1}{N_1} X N_2\ and\ from\ the\ figure\ above,\ this\ voltage\ is\ V_2$

$$Hence,\ \frac{V_1}{N_1} X N_2 = V_2$$

$$\Rightarrow \frac{V_2}{V_1} = \frac{N_2}{N_1} = Constant = K$$

As BC portion of the winding is considered as secondary, it can easily be understood that value of constant 'k' is nothing but turns ratio or voltage ratio of that **auto transformer**. When load is connected between secondary terminals i.e.between 'B' and 'C', load current I_2 starts flowing. The current in the secondary winding or common winding is the difference of I_2 and I_1.

Copper Savings in Auto Transformer

Now we will discuss the savings of copper in auto transformer compared to conventional two winding transformer.
We know that weight of copper of any winding depends upon its length and cross-sectional area. Again length of conductor in winding is proportional to its number of turns and cross-sectional area varies with rated current.
So weight of copper in winding is directly proportional to product of number of turns and rated current of the winding.

Therefore, weight of copper in the section AC proportional to,
$(N_1 - N_2)I_1$
and similarly, weight of copper in the section BC proportional to,
$N_2(I_2 - I_1)$
Hence, total weight of copper in the winding of auto transformer proportional to,
$(N_1 - N_2)I_1 + N_2(I_2 - I_1)$

$$\Rightarrow N_1I_1 - N_2I_1 + N_2I_2 - N_2I_1$$

$$\Rightarrow N_1I_1 + N_2I_2 - 2N_2I_1$$

$$\Rightarrow 2N_1I_1 - 2N_2I_1 (Since,\ N_1I_1 = N_2I_2)$$

$$\Rightarrow 2(N_1I_1 - N_2I_1)$$

In similar way it can be proved, the weight of copper in two winding transformer is proportional to,
$N_1I_1 - N_2I_2$

$$\Rightarrow 2N_1I_1 \qquad (Since,\ in\ a\ transformer\ N_1I_1 = N_2I_2)$$

$N_1I_1 + N_2I_2$

$\Rightarrow 2N_1I_1$ (Since, in a transformer $N_1I_1 = N_2I_2$)

Let's assume, W_a and W_{tw} are weight of copper in auto transformer and two winding transformer respectively,

$$Hence,\ \frac{W_a}{W_{tw}} = \frac{2(N_1I_1 - N_2I_1)}{2(N_1I_1)}$$

$$= \frac{N_1I_1 - N_2I_1}{N_1I_1} = 1 - \frac{N_2I_1}{N_1I_1}$$

$$= 1 - \frac{N_2}{N_1} = 1 - k$$

$$\therefore W_a = W_{tw}(1-k)$$

$$\Rightarrow W_a = W_{tw} - kW_{tw}$$

∴ Saving of copper in auto transformer compared to two winding transformer,

$$\Rightarrow W_{tw} - W_a = kW_{tw}$$

Auto transformer employs only single winding per phase as against two distinctly separate windings in a conventional transformer.

Advantages of using Auto Transformers

1. For transformation ratio = 2, the size of the **auto transformer** would be approximately 50% of the corresponding size of two winding transformer. For transformation ratio say 20 however the size would be 95 %. The saving in cost of the material is of course not in the same proportion. The saving of cost is appreciable when the ratio of transformer is low, that is lower than 2. Thus auto transformer is smaller in size and cheaper.

2. An auto transformer has higher efficiency than two winding transformer. This is because of less ohmic loss and core loss due to reduction of transformer material.

3. Auto transformer has better voltage regulation as voltage drop in resistance and reactance of the single winding is less.

Disadvantages of Using Auto Transformer

1. Because of electrical conductivity of the primary and secondary windings the lower voltage circuit is liable to be impressed upon by higher voltage. To avoid breakdown in the lower voltage circuit, it becomes necessary to design the low voltage circuit to withstand higher voltage.

2. The leakage flux between the primary and secondary windings is small and hence the impedance is low. This results into severer short circuit currents under fault conditions.

3. The connections on primary and secondary sides have necessarily needs to be same, except when using interconnected starring connections. This introduces complications due to changing primary and secondary phase angle particularly in the case of delta/delta connection.

4. Because of common neutral in a star/star connected auto transformer it is not possible to earth neutral of one side only. Both their sides should have their neutrality either earth or isolated.

5. It is more difficult to maintain the electromagnetic balance of the winding when voltage adjustment tappings are provided. It should be known that the provision of tapping on an auto transformer increases considerably the frame size of the transformer. If the range of tapping is very large, the advantages gained in initial cost is lost to a great event.

Applications of Auto Transformers

1. Compensating voltage drops by boosting supply voltage in distribution systems.

2. Auto transformers with a number of tapping are used for starting induction and synchronous motors.

3. **Auto transformer** is used as variac in laboratory or where continuous variable over broad ranges are required.

INSTRUMENT TRANSFORMER

Instrument Transformers are used in AC system for measurement of electrical quantities i.e. voltage, current, power, energy, power factor, frequency. **Instrument transformers** are also used with protective relays for protection of power system.

Basic function of **Instrument transformers** is to step down the AC System voltage and current. The voltage and current level of power system is very high. It is very difficult and costly to design the measuring instruments for measurement of such high level voltage and current. Generally measuring instruments are designed for 5 A and 110 V.

The measurement of such very large electrical quantities, can be made possible by using the Instrument transformers with these small rating measuring instruments. Therefore these instrument transformers are very popular in modern power system.

Advantages of Instrument Transformers

1. The large voltage and current of AC Power system can be measured by using small rating measuring instrument i.e. 5 A, 110 – 120 V.
2. By using the instrument transformers, measuring instruments can be standardized. Which results in reduction of cost of measuring instruments. More ever the damaged measuring instruments can be replaced easy with healthy standardized measuring instruments.

3. Instrument transformers provide electrical isolation between high voltage power circuit and measuring instruments. Which reduces the electrical insulation requirement for measuring instruments and protective circuits and also assures the safety of operators.

4. Several measuring instruments can be connected through a single transformer to power system.

5. Due to low voltage and current level in measuring and protective circuit, there is low power consumption in measuring and protective circuits.

Types of Instrument Transformers

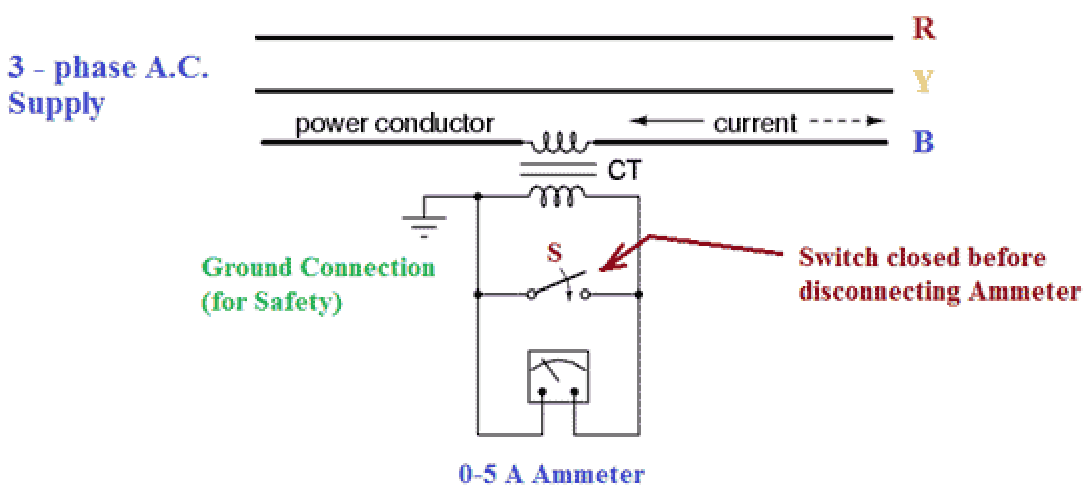
Instrument transformers are of two types –

1. Current Transformer (C.T.)

2. Potential Transformer (P.T.)

Current Transformer (C.T.)

Current transformer is used to step down the current of power system to a lower level to make it feasible to be measured by small rating Ammeter (i.e. 5A ammeter). A typical connection diagram of a current transformer is shown in figure below.



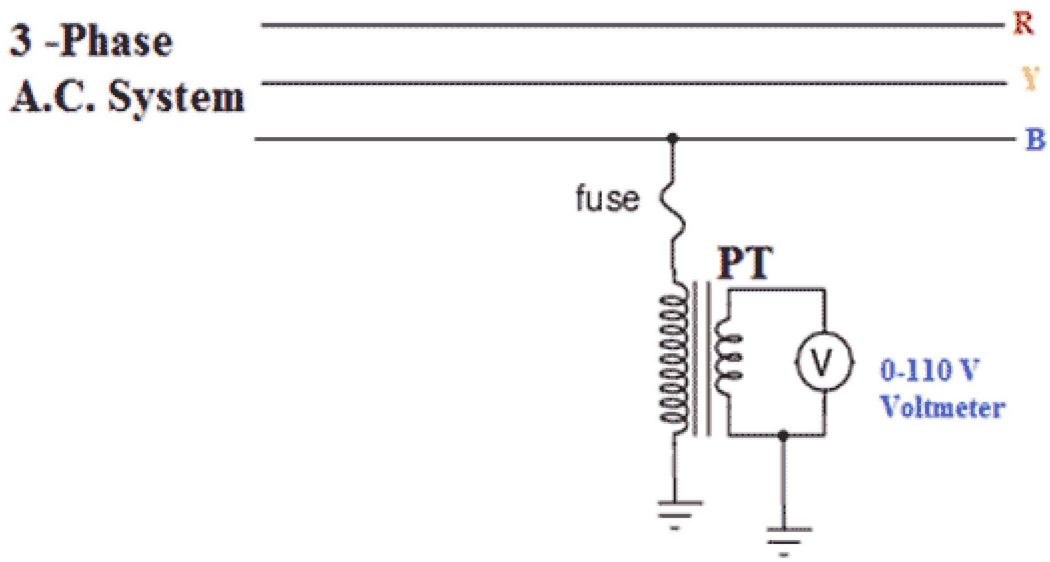
Current Transformer (C.T.)

Primary of C.T. is having very few turns. Sometimes bar primary is also used. Primary is connected in series with the power circuit. Therefore, sometimes it also called **series transformer**. The secondary is having large no. of turns. Secondary is connected directly to an ammeter. As the ammeter is having very small resistance.

Hence, the secondary of current transformer operates almost in short circuited condition. One terminal of secondary is earthed to avoid the large voltage on secondary with respect to earth. Which in turns reduce the chances of insulation breakdown and also protect the operator against high voltage. More ever before disconnecting the ammeter, secondary is short circuited through a switch 'S' as shown in figure above to avoid the high voltage build up across the secondary.

Potential Transformer (P.T.)

Potential transformer is used to step down the voltage of power system to a lower level to make is feasible to be measured by small rating voltmeter i.e. 110 – 120 V voltmeter. A typical connection diagram of a potential transformer is showing figure below.



Potential Transformer (P.T.)

Primary of P.T. is having large no. of turns. Primary is connected across the line (generally between on line and earth). Hence, sometimes it is also called the **parallel transformer**. Secondary of P.T. is having few turns and connected directly to a voltmeter. As the voltmeter is having large resistance. Hence the secondary of a P.T. operates almost in open circuited condition. One terminal of secondary of P.T. is earthed to maintain the secondary voltage with respect to earth. Which assures the safety of operators.

Difference between C.T. and P.T.

Few differences between C.T. and P.T. are listed below –

Sl. No.	Current Transformer (C.T.)	Potential Transformer (P.T.)
1	Connected in series with power circuit.	Connected in Parallel with Power circuit.
2	Secondary is connected to Ammeter.	Secondary is connected to Voltmeter.
3	Secondary works almost in short circuited condition.	Secondary works almost in open circuited condition.
4	Primary current depends on power circuit current.	Primary current depends on secondary burden.
5	Primary current and excitation vary over wide range with change of power circuit current	Primary current and excitation variation are restricted to a small range.
6	One terminal of secondary is earthed to avoid the insulation break down.	One terminal of secondary can be earthed for Safety.
7	Secondary is never be open circuited.	Secondary can be used in open circuit condition.

GANDHI INSTITUTE OF EXCELLENT TECHNOCRATS BHUBANESWAR

DEPARTMENT OF ELECTRICALENGINEERING

LECTURE NOTES

Year & Semester: 2^{ND} Year IV Semester Electrical

Subject Code/Name: TH-2, ANALOG ELECTRONICS AND OP-AMP

Prepared By:-

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Subtractor
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Differentiator
Comparator

COURSE OUTCOME (CO):

After the completion of the course the students are able to:

1. Understand the working and current voltage characteristic of semiconductor devices.
2. Analyze different parameters of various rectifier circuits.
3. Explain the working of different types of filter circuit.
4. Develop the ability to understand the working of different BJT and FET circuits.
5. Describe different method of biasing of BJT and FET.
6. Analyze DC circuits and relate AC models with their physical operation.
7. Explain the working and frequency response of different types of transistor amplifier.
8. Develop knowledge of operational amplifiers and their application.
9. Observe the effect of feedback circuits and explain the working of different types of oscillators.

UNIT-1: PN JUNCTION DIODE

**SEMICONDU
CTOR:**

Semiconductors (*e.g. germanium, silicon etc*.) are those substances whose electrical conductivity lies in between conductors and insulators. In terms of energy band, the valence band is almost filled and conduction band is almost empty. Further, the energy gap between valence and conduction bands is very small. The semiconductor has:

- Filled valence band
- Empty conduction band
- Small energy gap or forbidden gap (1 ev) between valence and conduction bands.
- Semiconductor virtually behaves as an insulator at low temperatures. However, even at room temperature, some electrons cross over to the conduction band, imparting little conductivity (i.e. conductor).

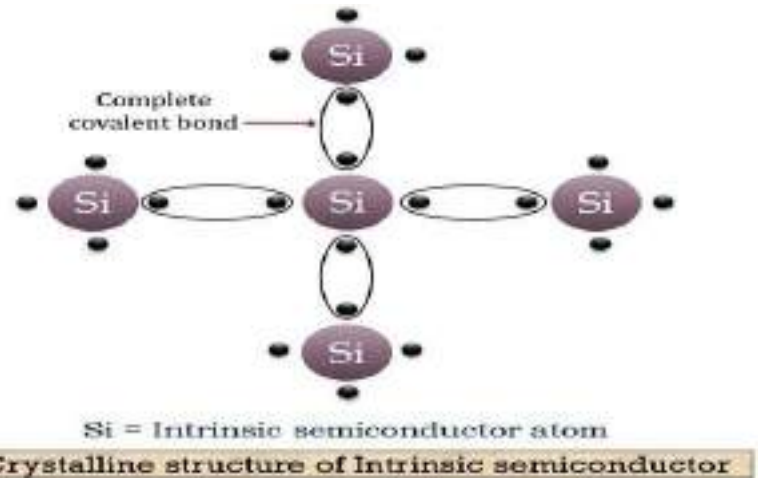
Types of semiconductors:-

Semiconductors are classified into two types:-

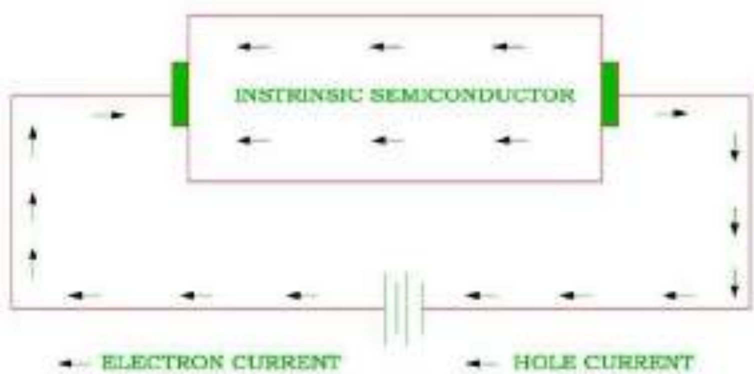
- ❖ Intrinsic semiconductors
- ❖ Extrinsic semiconductors
- ➢ Extrinsic semiconductors are also of two types:-
- ▪ P-type semiconductors
- ▪ N-type semiconductors

Intrinsic semiconductors

- A semiconductor in an extremely pure form is known as anintrinsic semiconductor. When electric field is applied across an intrinsic semiconductor, the current conduction takes place by two processes i.e. by free electrons and holes.
- The free electrons are produced due to the breaking up of some covalent bonds by thermal energy. At the same time, holes are created in the covalent bonds.
- Under the influence of electric field, conduction through the semiconductor is by both free electrons and holes. Therefore, the total current inside the semiconductor is the sum of currents due to free electrons and holes.



Crystalline structure of Intrinsic semiconductor



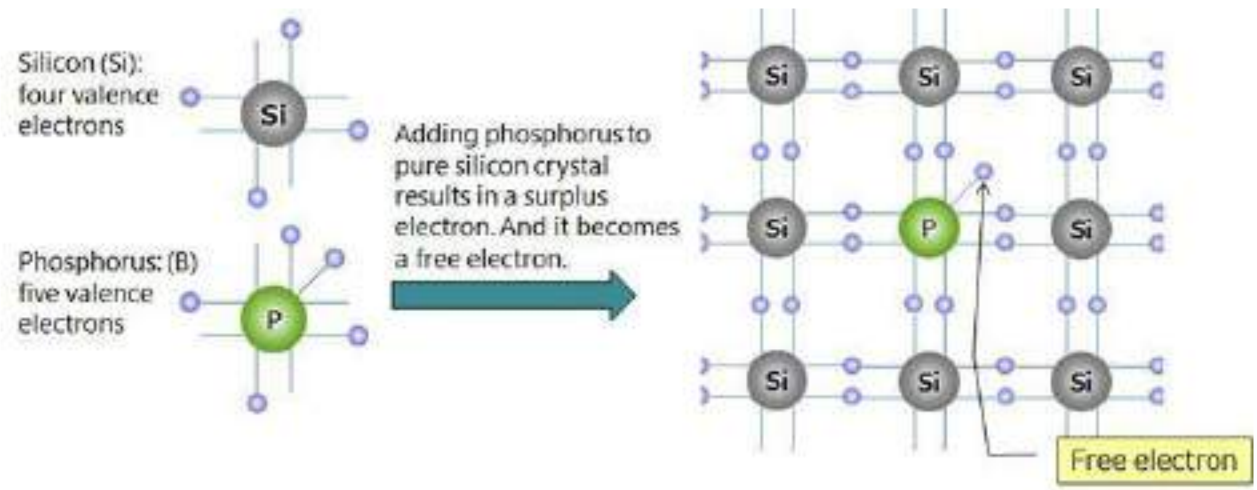
Extrinsic semiconductors

An extrinsic semiconductor is a semiconductor doped by addition of small amount impurity. The process of adding impurities to a semiconductor is known as doping. The purpose of adding impurity is to increase either the number of free electrons or holes in the semiconductor crystal. Depending upon the type of impurity added, extrinsic semiconductors are classified into:

- n-type semiconductor
- p-type semiconductor

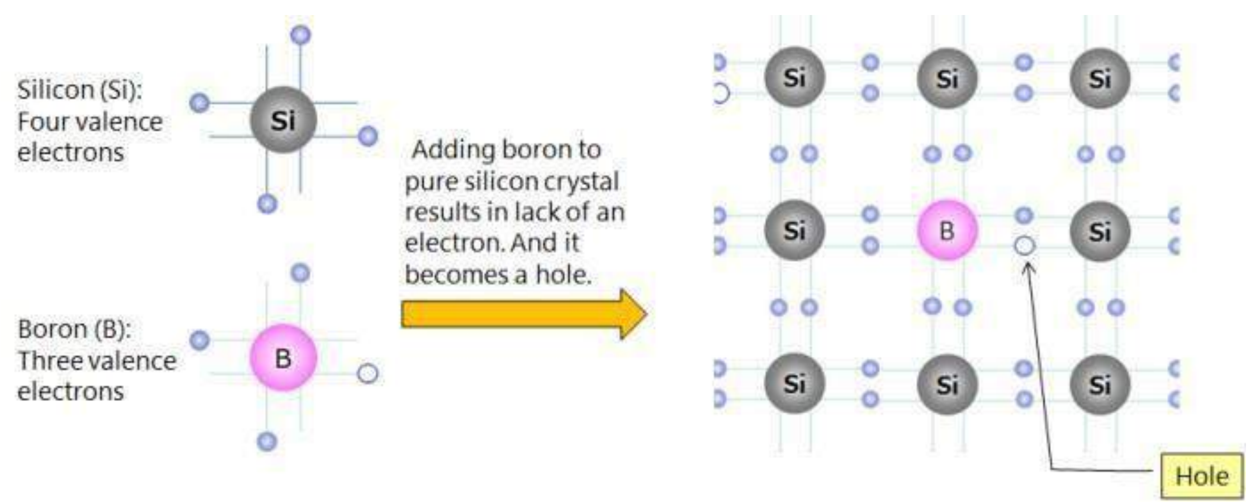
N-type semiconductors

When a small amount of pentavalent impurity is added to a pure semiconductor, it is known as n-type semiconductor. Examples of pentavalent impurities are arsenic, antimony, Bismuth and Phosphorous etc. Such impurities are known as donor impurities because they donate or provide free electrons to the semiconductor crystal.Electrons are said to be the majority carriers whereas holes are the minority carriers.



P-type semiconductors

When a small amount of trivalent impurity is added to a pure semiconductor, it is called p-type Semiconductor. Examples of trivalent impurities are gallium, indium, boron etc. The addition of trivalent impurity provides a large number of holes in the semiconductor. Such impurities are known as acceptor impurities because the holes created can accept the electrons. In a p type semiconductor holes are the majority carriers and electrons are the minority carriers.



PN JUNCTION:-

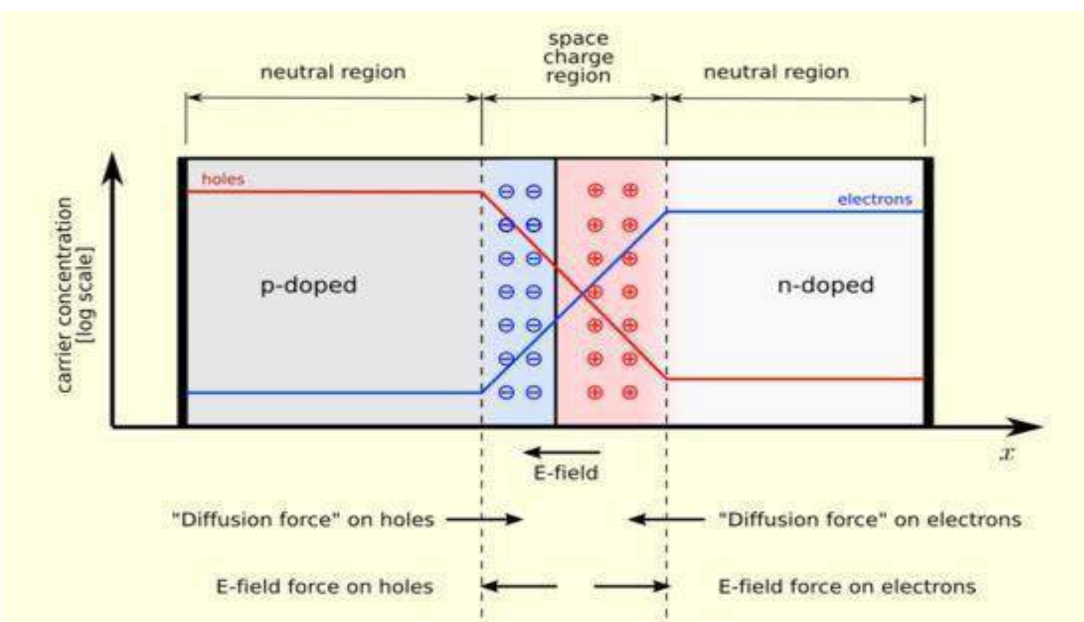
Zero Biased Junction Diode

When a diode is connected in a Zero Bias condition, no external potential energy is applied to the PN junction. However if the diodes terminals are shorted together, a few holes (majority carriers) in the P-type material with enough energy to overcome the potential barrier will move across the junction against this barrier potential. This is known as the "Forward Current" and is referenced as I_F. Likewise, holes generated in the N-type material (minority carriers), find this situation favorable and move across the junction in the opposite direction. This is known as the "Reverse Current" and is referenced as I_R. This transfer of electrons and holes back and forth across the PN junction is known as diffusion.

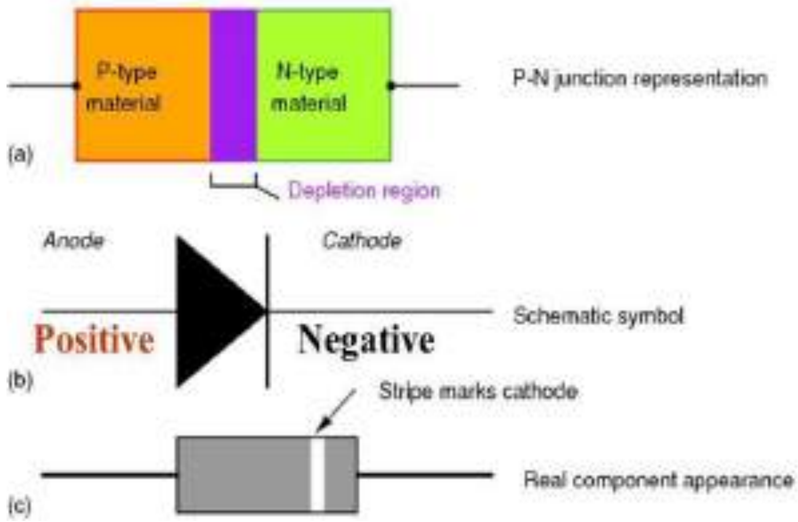
The potential barrier that now exists discourages the diffusion of any more majority carriers across the junction. However, the potential barrier helps minority carriers (few free electrons in the P-region and few holes in the N-region) to drift across the junction.

Then an "Equilibrium" or balance will be established when the majority carriers are equal and both moving in opposite directions, so that the net result is zero current flowing in the circuit. When this occurs the junction is said to be in a state of "Dynamic Equilibrium".

The minority carriers are constantly generated due to thermal energy so this state of equilibrium can be broken by raising the temperature of the PN junction causing an increase in the generation of minority carriers, thereby resulting in an increase in leakage current but an electric current cannot flow since no circuit has been connected to the PN junction.

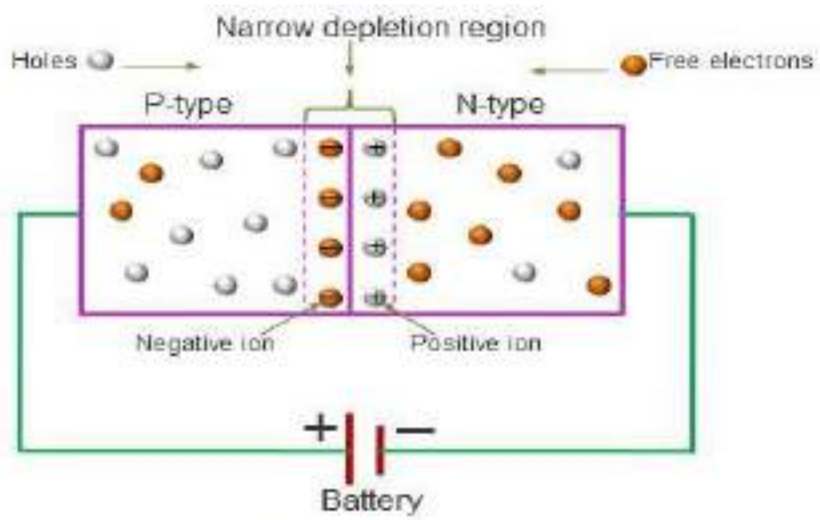


Diode Symbol



FORWARD BIASING:-

- A P-N junction diode is said to be forward biased when the positive terminal of a cell or battery is connected to the p-side of the junction and the negative terminal to the *n* side.
- When diode is forward-biased the depletion region narrows and consequently, the potential barrier is lowered.
- This causes the majority charge carriers of each region to cross into the other region.
- The electrons travel from the n-side to the p-side and go to the positive terminal of the battery.
- The holes that travel from the p-side to the n-side combine with the electrons injected into the n-region from the negative terminal of the battery.
  With the increase in forward bias voltage, the depletion region eventually becomes thin, thus reducing electrical resistance.
- This causes electrons to pass through the junction resulting in an exponential rise in current.
- This way the diode conducts when forward-biased.

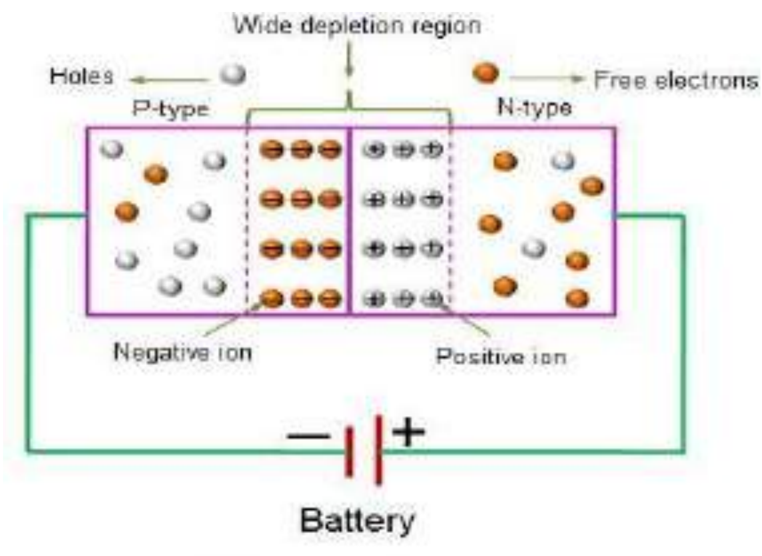


Forward bias

REVERSE BIASING:-

- A pn-junction diode is said to be reverse biased when the positive terminal of a cell or battery is connected to the n-side of the junction and the negative terminal to the p-side.
- When reverse biased, the depletion region widens and the potential barrier is increased.
- The polarity of the battery extracts the majority charge carriers of each region.
- The holes in the p-region from the electrons injected into the p-region from the negative terminal of the battery.

- The electrons in the n-region go to the positive terminal of the battery.
- This way, the majority charge carrier concentration in each region decreases against the equilibrium values and the reverse biased junction diode has a high resistance.
- Thus, the diffusion current across the junction becomes zero.
- Thus, the diode does not conduct when reverse biased and is said to be in a quiescent or non-conducting state i.e., it acts as an open switch.



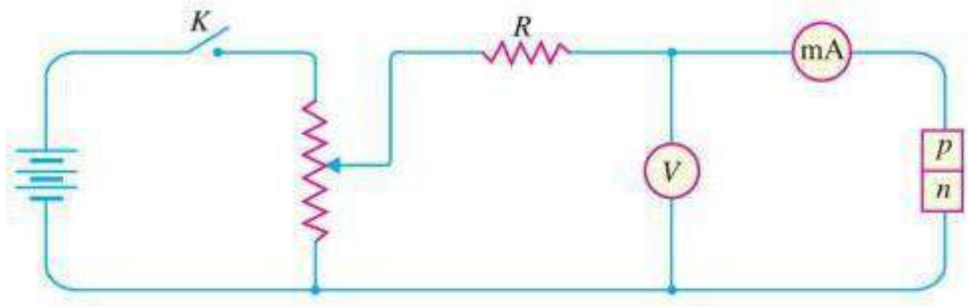
Reverse bias

VI CHARACTERISTICS OF PN JUNCTION:-

Volt-ampere (V-I) characteristics of a pn junction or semiconductor diode is the curve between voltage across the junction and the current through the circuit.

Normally the voltage is taken along the x-axis and current along y-axis.

The circuit connection for determining the V-I characteristics of a pn junction is shown in the figure below.



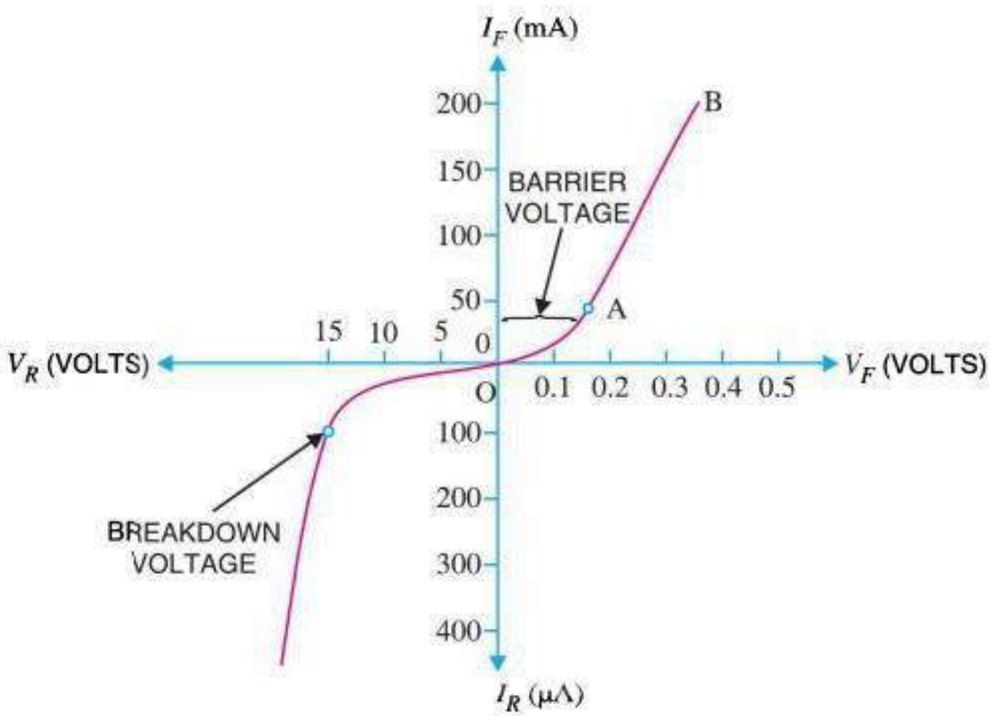
The characteristics can be explained under three cases, such as :

1. Zero bias
2. Forward bias
3. Reverse bias

Case-1: Zero Bias

In zero bias condition, no external voltage is applied to the pn junction i.e. the circuit is open at K.

Hence, the potential barrier at the junction does not permit current flow.
Therefore, the circuit current is zero at V=0 V, as indicated by point O in figure below.



Case-2: Forward Bias

In forward biased condition, p-type of the pn junction is connected to the positive terminal and n-type is connected to the negative terminal of the external voltage.

This results in reduced potential barrier.

At some forward voltage i.e. 0.7 V for Si and 0.3 V for Ge, the potential barrier is almost eliminated and the current starts flowing in the circuit.

Form this point, the current increases with the increase in forward voltage. Hence a curve OB is obtained with forward bias as shown in figure above.

From the forward characteristics, it can be noted that at first i.e. region OA, the current increases very slowly and the curve is non-linear. It is because in this region the external voltage applied to the pn junction is used in overcoming the potential barrier.

However, once the external voltage exceeds the potential barrier voltage, the potential barrier is eliminated and the pn junction behaves as an ordinary conductor. Hence, the curve AB rises very sharply with the increase in external voltage and the curve is almost linear.

Case-3: Reverse Bias

In reverse bias condition, the p-type of the pn junction is connected to the negative terminal and n-type is connected to the positive terminal of the external voltage.

This results in increased potential barrier at the junction.

Hence, the junction resistance becomes very high and as a result practically no current flows through the circuit.

However, a very small current of the order of μA, flows through the circuit in practice. This is knows as reverse saturation current (I_S) and it is due to the minority carriers in the junction.
As we already know, there are few free electrons in p-type material and few holes in n-type material. These free electrons in p-type and holes in n-type are called minority carriers.

The reverse bias applied to the pn junction acts as forward bias to their minority carriers and hence, small current flows in the reverse direction.

If the applied reverse voltage is increased continuously, the kinetic energy of the minority carriers may become high enough to knock out electrons from the semiconductor atom.

At this stage breakdown of the junction may occur. This is characterized by a sudden increase of reverse current and a sudden fall of the resistance of barrier region. This may destroy the junction permanently.

DC LOAD LINE-

A circuit supplied with dc power as the external source of the circuit. There exist both alternating and direct currents in the circuit. The reactive components of the circuits are made zero and the straight line is drawn above the voltage-current characteristics curves. Hence these results in the formation of intersecting point referred to an operating point. The straight that is drawn for this purpose is defined as the DC load line.

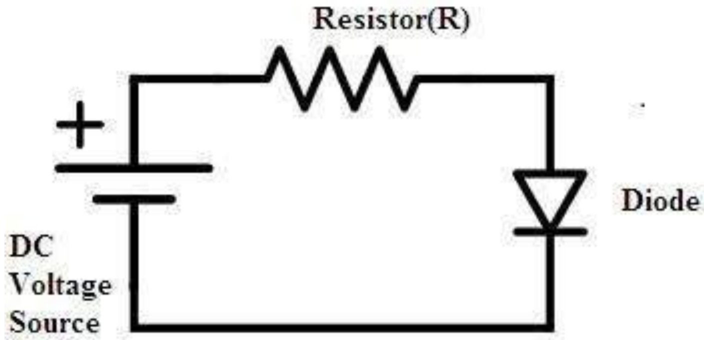
DC Load Line of a Diode and Its Equation

DC load line for a non-linear device is drawn by making the reactive components as zero. Hence a diode is considered as a non-linear device and its voltage and current characteristics are exponentially related to each other. The formation of the intersection point for the characteristic curve and the straight line or dc-load line can be analyzed better by considering the example for the diode as in forward bias condition.

Let us consider a diode connected to the resistor(R), source of voltage (V_{DD}) in series. The diode is connected in forward bias so that the forward current and the forward voltages flowing through the circuit. As per the Kirchhoff's current law, the current flowing through the diode (I_D) and the resistor (I_R) is equal.

$$\mathbf{I_D = I_R}$$

Analysis of the circuit is done by applying Kirchhoff's voltage law (KVL). This law results in the formation of the final equation for the dc load line. Here the dc voltage is the biasing voltage of the circuit by keeping any further reactive components as zero.



Diode-operating-in-forward-bias-for-the-analysis-of-dc-load

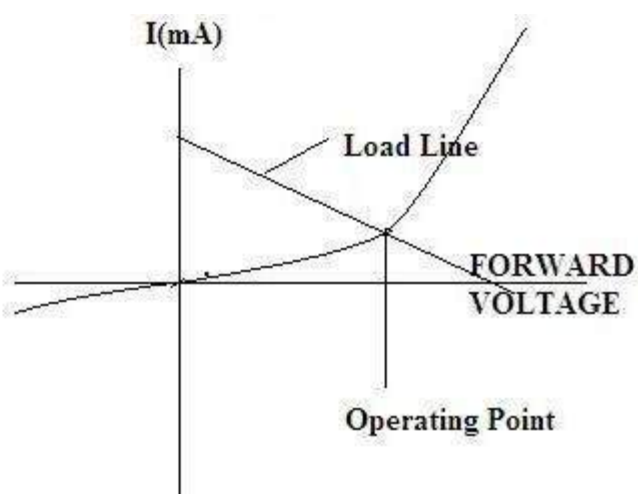
Once the Kirchhoff's voltage law is applied to the circuit an equation is obtained for voltages and currents in the circuit.

$$V_{DD} = V_D + I_D R$$

$$V_D = I_D R - V_{DD}$$

Where V_{DD}, is the applied dc source voltage and V_D is the voltage across the diode. Hence the above can be considered as the equation for the diode. The voltage and current characteristics of the diode in forward bias condition can be drawn. By our previous analysis on the condition of the diode in forward bias applied a voltage and the generated current in the circuit are exponentially related to each other.

After a certain cut-off voltage, the diode starts operating in forward bias condition. To this slope, the technique is considered and a straight line on the v-i characteristics is drawn. The slope here for the above general circuit for the diode is V_{DD}/R.



Dc-load-line-and-the-formation-of-operating-point

In this way, the analysis for the dc-load line is done for the non-linear device like a diode. Depending on the type of non-linear device some part of the analysis differs but the technique remains the same. This type of method comes under the graphical analysis because here the characteristics curve is considered for the formation of the dc-load line.

IMPORTANT TERMS SUCH AS IDEAL DIODE, KNEE VOLTAGE

Ideal Diode:

An ideal diode is one kind of an electrical component that performs like an ideal conductor when voltage is applied in forward bias and like an ideal insulator when the voltage is applied in reverse bias. So when +ve voltage is applied across the anode toward the cathode, the diode performs forward current immediately. When a voltage is applied in reverse bias, then ir performs no current at all. This diode operates like a switch. When the diode is in forward bias, it works like a closed switch. Whereas, if an ideal diode is in reverse bias, then it works like an open switch.

Knee Voltage:

The forward voltage at which the current through the junction starts increasing rapidly, is called knee voltage or cut-in voltage. It is generally 0.6v for a diode.

JUNCTIONS BREAKDOWN:

- ✓ The Avalanche Breakdown and Zener Breakdown are two different mechanisms by which a PN junction breaks.
- ✓ The Zener and Avalanche breakdown both occur in diode under reverse bias.
- ✓ The avalanche breakdown occurs because of the ionization of electrons and hole pairs whereas the Zener diode occurs because of heavy doping.

AVALANCHE BREAKDOWN:

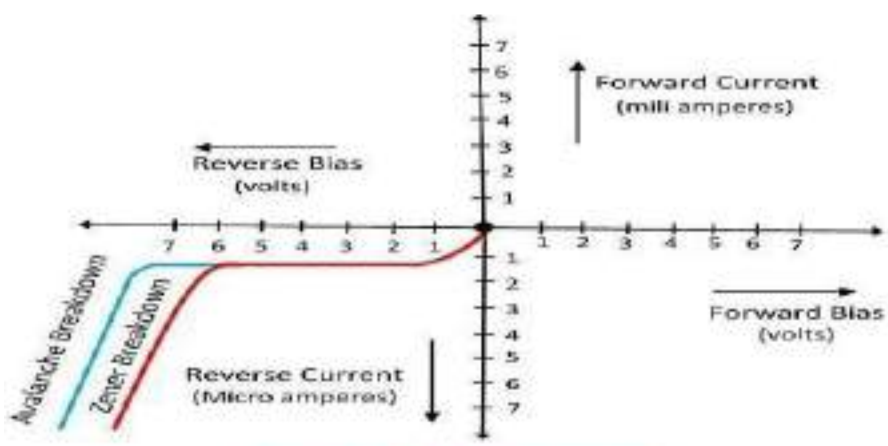
The mechanism of avalanche breakdown occurs because of the reverse saturation current. The P-type and N-type material together forms the PN-junction. The depletion region develops at the junction where the P and N-type material contact. The P and N-type materials of the PN junction are not perfect, and they have some impurities in it, i.e., the p-type material has some electrons, and the N-type material has some hole in it. The width of the depletion region varies. Their width depends on the bias applied to the terminal of P and N region. The reverse bias increases the electrical field across the depletion region. When the high electric field exists across the depletion, the velocity of minority charge carrier crossing the depletion region increases. These carriers collide with the atoms of the crystal. Because of the violent collision, the charge carrier takes out the electrons from the atom. The collision increases the electron-hole pair. As the electron-hole induces in the high electric field, they are quickly separated and collide with the other atoms of the crystals. The process is continuous, and the electric field becomes so much higher than the reverse current starts flowing in the PN junction. The process is known as the Avalanche breakdown. After the breakdown, the junction cannot regain its original position because the diode is completely burnt off.

ZENER BREAKDOWN:

The PN junction is formed by the combination of the p-type and the n-type semiconductor material. The combination of the P-type and N-type regions creates the depletion region. The width of the depletion region depends on the doping of the P and N-type semiconductor material. If the material is heavily doped, the width of the depletion region becomes very thin. The phenomenon of the Zener breakdown occurs in the very thin depletion region. The thin depletion region has more numbers of free electrons. The reverse bias applied across the PN junction develops the electric field intensity across the depletion region. The strength of the electric field intensity becomes very high. The electric field intensity increases the kinetic energy of the free charge carriers. Thus the carrier starts jumping from one region to another. These energetic charge carriers collide with the atoms of the p-type and n-type material and produce the electron-hole pairs. The reverse current starts flowing in the junction because of which depletion region becomes entirely vanishes. This process is known as the Zener breakdown. In Zener breakdown, the junction is not completely damaged. The depletion region regains it original position after the removal of the reverse voltage.The voltage of Zener breakdown is less than the Avalanche breakdown.

BreakdownCharacteristicGraph

The graphical representation of the Avalanche and Zener breakdown is shown in the figure below.



Breakdown Characteristic

P-NDIODECLIPPERCIRCUIT:

Clipper circuits are the circuits that clip off or remove a portion of an input signal, without causing any distortion to the remaining part of the waveform. These are also known as clippers, clipping circuits, limiters, slicers etc.
Clippers are basically wave shaping circuits that control the shape of an output waveform. It consists of linear and non-linear elements but does not contain energy storing elements.

The basic operation of a diode clipping circuits is such that, in forward biased condition, the diode allows current to pass through it, clamping the voltage. But in reverse biased condition, no any current flows through the diode, and thus voltage remains unaffected across its terminals.

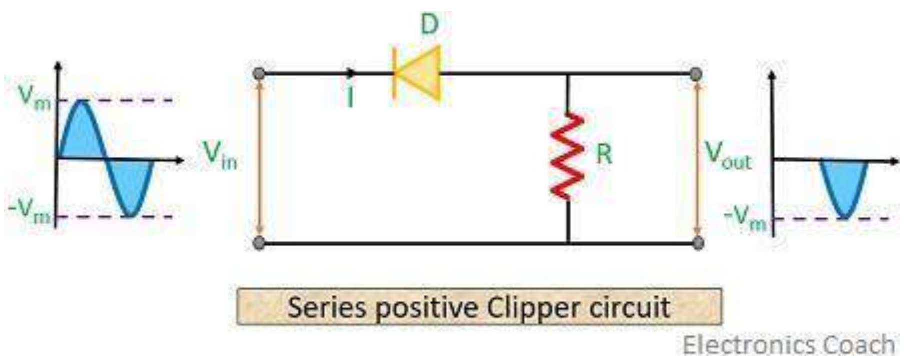
Clipper circuits are basically termed as protection devices. As electronic devices are voltage sensitive and voltage of large amplitude can permanently destroy the device. So, in order to protect the device clipper circuits are used.

Usually, clippers employ resistor–diode combination in its circuitry.

Classification of Clipper circuits

Clippers are basically classified in the following categories:

- Series positive Clipper circuit



Series positive Clipper circuit

Let's have a look at the circuit diagram of a series positive clipper. Here, the diode is connected in series with the output thus it is named so.

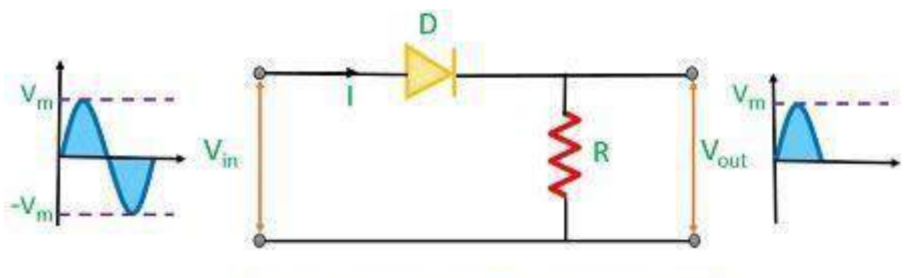
The positive half of the input waveform reverse biases the diode. Thus it acts as an open switch and all the applied input voltage drops across the diode. Resultantly providing no output voltage for positive half of the input waveform.

For the negative half of the input waveform, the diode is in the forward biased state. Thus it acts as a closed switch causing no any voltage drop at the diode.

Hence input voltage will appear across the resistor, ultimately at the output of the circuit.

- Series negative clipper circuit

The circuit below shows the figure of a series negative clipping circuit.



Series negative Clipper circuit

Electronics Coach

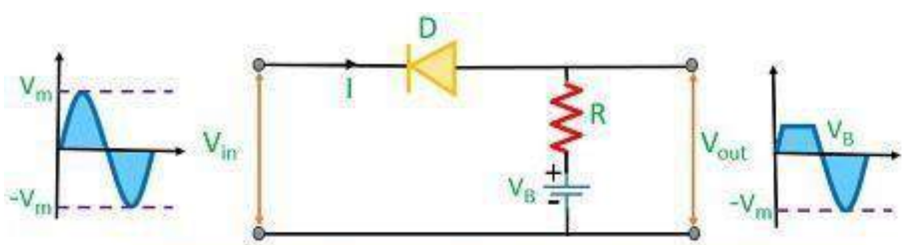
Here, during the positive half cycle of input waveform, the diode becomes forward biased, thus ensuring a closed circuit. Due to which current appears across the resistor of the circuit.

For negative half of the input waveform, the diode now becomes reverse biased acting as an open switch. This causes no current to flow through the circuit. Resultantly providing no output for negative half of the input waveform.

- Series positive clipper circuit with bias

Whenever there is a need to clip or remove a certain portion of positive half of input waveform, series positive clippers with biasing are needed.

1. The case of a Positively biased circuit:



Series positive Clipper circuit with positive biasing

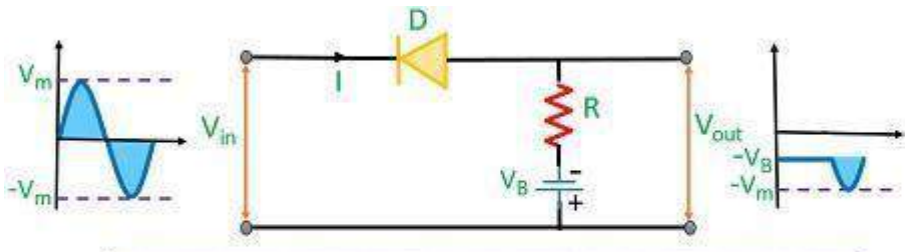
Electronics Coach

Here in the circuit shown above, we can see that the diode is in forward bias condition concerning the battery. But positive half of the input waveform puts the diode in reverse biased condition.

The diode will conduct until the supply voltage is less than the battery potential. As battery potential dominates the supply voltage, the signal appears at the positive half of output waveform. But as the supply voltage exceeds the battery potential, the diode is now reverse biased. Resultantly no further current will flow through the diode.

For the negative half cycle of the input waveform, the diode is forward biased concerning both supply voltage and battery potential. Hence, we achieve a complete negative half cycle at the output waveform.

2. The case of a negatively biased circuit:



Series positive Clipper circuit with negative biasing

As we can see in the circuit shown above, the diode is reverse bias due to both supply voltage and battery potential. This cuts off the complete positive half of the input waveform.

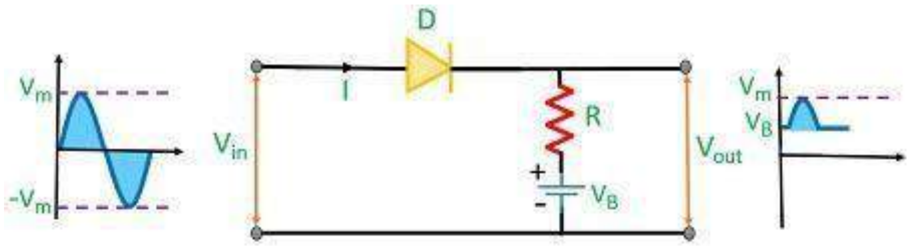
But during the negative half cycle of the input waveform, the diode is in forward biased condition due to supply voltage but is reverse biased by the battery potential.

Here also initially when battery dominates the supply voltage, the diode is in reverse biased condition. But, as the supply voltage becomes greater than the battery potential, the diode will automatically come in forward biased condition. Thus, the signal starts to appear at the output.

- Series negative clipper circuits with bias

  Whenever there is a need to clip or remove a certain portion of the negative half of the input waveform, then series negative clipper circuits with biasing is needed.

  1. The case of a positively biased circuit:



Series negative Clipper circuit with positive biasing

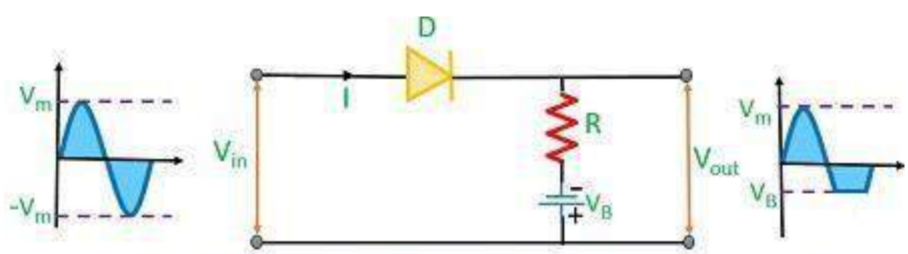
Electronics Coach

Here in the circuit, we have noticed that for the positive half of input waveform, the diode becomes forward bias. However, the battery potential causes the diode to be at the reverse biased condition.

In the beginning, the battery potential is higher than the supply voltage. Thus causing no current to flow through the circuit at that time. But when the positive half supply voltage exceeds the battery potential then diode becomes forward biased and starts conduction. Hence signal appears at the output.

The negative half cycle of the input waveform causes the diode to be reverse biased. At the same time applied battery potential reverse biases the diode. Thus no any signal is achieved at the output in such a condition.

2. The case of negatively biased circuits:



Series negative Clipper circuit with negative biasing

Electronics Coach

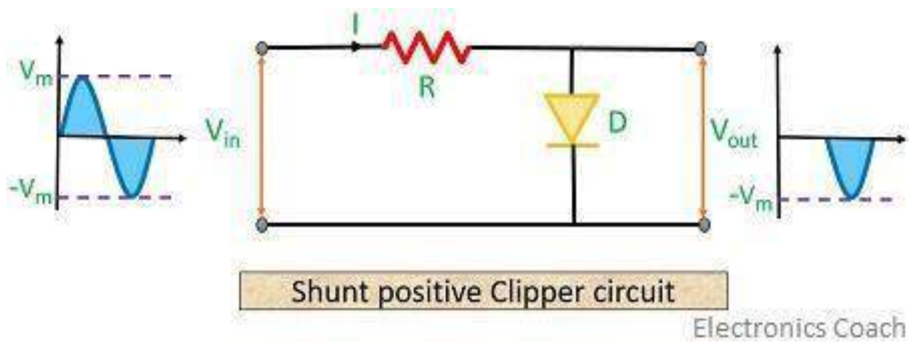
For the positive half of the input cycle, supply voltage and battery potential both cause the diode to be in forward biased condition. Thus complete output for the positive half is achieved in this case.

But in the case of the negative half cycle of the input waveform, the diode is now in reverse biased condition. At the same time due to battery potential, the diode is forward biased. So, until the battery potential is greater than the supply voltage, current flows through the circuit. Hence output is achieved.

But as the supply voltage becomes greater than the battery potential, the diode will now become reverse biased and hence conduction stops. Ultimately no any signal appears for that portion of the input waveform at the output.

- Shunt positive Clipper circuits

  Now, let's have a look at the circuit of shunt clippers shown below



Shunt positive Clipper circuit

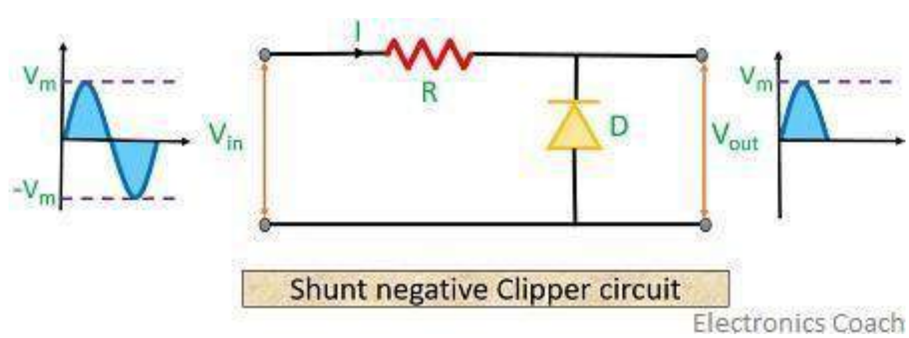
Here, the diode is connected in parallel with the load. Hence its working principle is exactly opposite to that of the shunt clippers. In shunt positive clippers, the output signal is observed only when the diode is reverse biased.

During the positive half of the input signal, the diode gets forward biased due to which the current flows through the diode. Hence, due to the parallel combination of diode and load, no current is observed at the load. Resultantly no output waveform for the positive half is achieved.

On the contrary, during the negative half of the input signal, the diode gets reverse biased. Thus no current flows through it, and the output current is observed at the load. So, for the negative half of input, the entire negative half appears at the output.

- Shunt negative Clippers circuits

  Let's have a look at the figure shown below of shunt negative clippers.



Shunt negative Clipper circuit

For negative shunt clippers, during the positive half of input, the diode gets reverse biased. Thus no current flows through it, and the output current is observed at the load.

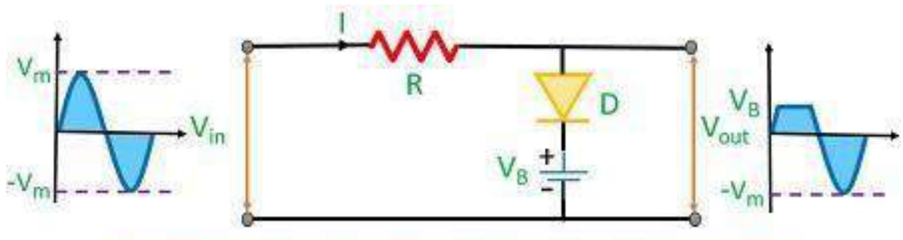
Hence output signal is achieved for positive half of the input signal.

During the negative half of the input signal, the diode gets forward biased and hence no load current is achieved. Ultimately no output is observed for negative half of the input signal.

- Shunt positive clipper circuits with bias

Here, we will discuss both the cases of positive and negative biasing separately as we have done in series clipper with bias.

1. The case of positively biased circuits:



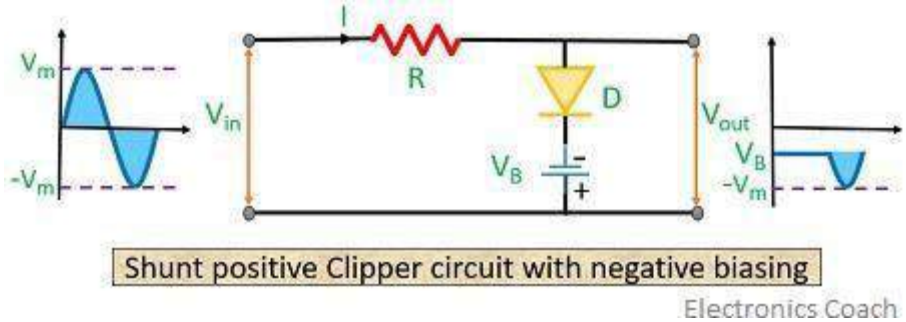
Shunt positive Clipper circuit with positive biasing

During the positive half of the input cycle, the diode gets forward biased but at the same time battery potential causes the diode to be at the reverse biased condition.

Until the battery potential is greater than the supply voltage, the reverse biased diode by the battery potential does not allow the flow of current through it. Hence current flows through load and signal is observed at the output. But when input voltage increases more than the battery potential, the diode becomes forward biased due to the supply voltage. Hence, allowing current to flow through the diode and no output is achieved.

In case of negative half of the input cycle, both battery potential and supply voltage reverse biases the diode. Hence we get a complete negative half cycle at the output.

2. The case of negatively biased circuits:



In this case, the negative half of the input cycle and the applied battery potential both causes the diode to be forward biased. Thus current flows through the diode, and no signal is observed across the output.

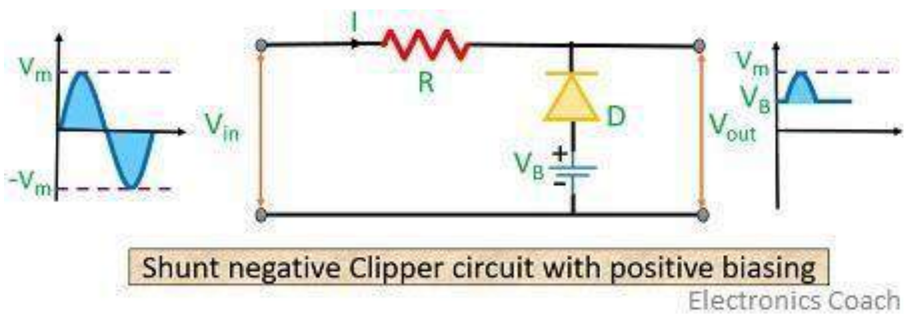
During the negative half of the input cycle, the diode is reverse biased due to supply voltage but is in the forward biased state due to battery potential.

So, here output will not appear at the load until battery potential is higher than the supply voltage. As the supply voltage exceeds the battery potential, the diode stops conduction due to the reverse bias condition. Hence signal appears at the output.

- Shunt negative clipper circuits with bias

  Now, let's move further and discuss the case of positive and negative biasing separately.

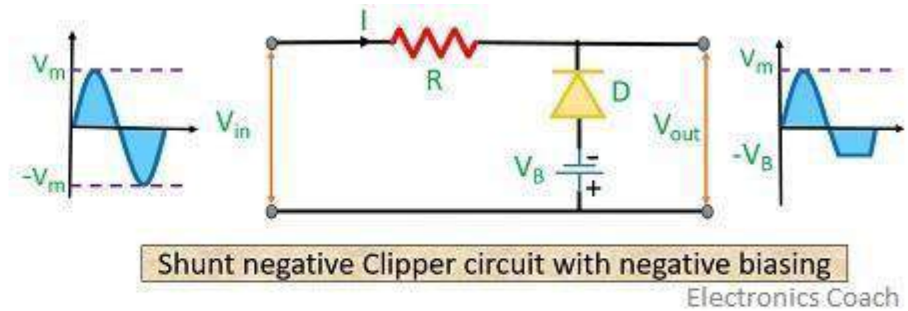
  1. The case of positively biased circuits:



In case of positive half of the input signal, the diode gets reverse biased. However, the battery potential causes the diode to be in forward biased condition. So, until the battery potential is higher than the supply voltage, the diode conducts due to forward bias. But as the supply voltage exceeds the battery potential the conduction through diode stops. Ultimately signal appears at the output.

During the negative half of the input cycle, the diode gets forward biased because of battery potential and supply voltage. Hence, no output signal appears for the negative half of the input cycle.

2. The case of negatively biased circuits:



During the positive half of the input cycle, the diode gets reverse biased due to both battery potential and supply voltage. Resultantly, we have complete positive half of input signal at the output.

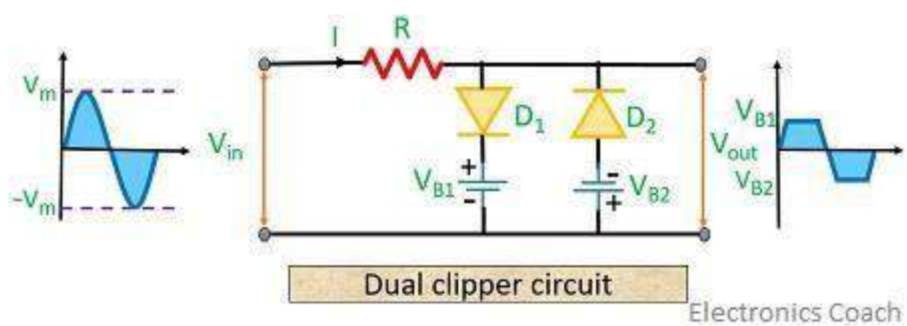
During the negative half of the input cycle, the diode gets forward biased due to supply voltage but is reverse biased by the battery potential. Thus, the output signal is achieved until the battery potential is higher than the supply voltage.

As the supply voltage exceeds the battery potential, the diode completely starts conduction and signal does not appear at the output.

- Dual clipper circuits

Whenever there is a need to remove a certain portion of both positive and negative half of the signal, then dual clipper circuits are used.

Let's have a look at the circuit of dual clipper shown below-



During positive half of the input cycle, diode D_1 gets forward biased due to supply voltage but is in reverse bias state due to battery potential V_{B1}. At the same time, diode D_2 is in reverse bias due to both supply voltage and battery potential V_{B2}.

Until battery voltage exceeds the supply voltage D_1 will be in reverse-biased state and D_2 is already in reverse bias condition. Hence signal is achieved at the output. But as the supply voltage exceeds the battery potential, it causes diode D_1 to be forward biased. Hence, no any further signal for the positive half is achieved in this case.

During the negative half of the input cycle, diode D_1 will be reverse biased due to both supply voltage and battery potential.

Conversely, diode D_2 will be forward biased by the supply voltage but is reverse biased by the battery potential V_{B2}.

Until the supply voltage is less than the battery potential, D_2 will be in reverse-biased state and signal appears at the output in this condition. But, as the supply voltage exceeds the battery potential, D_2 will be forward biased. Hence no any signal is obtained at the output.

Applications of Clipper circuits

These are used in numerous applications such as in transmission, in overvoltage protection, in the modification of input waveform, in voltage limiting etc.

P-N DIODECLAMPER CIRCUIT:

Clamper circuits are the electronic circuits that shift the dc level of the AC signal. Clampers are also known as DC voltage restorers or level shifter. Clampers are basically classified as positive and negative that includes both biased and unbiased conditions individually. These circuits are used to clamp an input signal to a different dc level. It basically adds dc component to the applied input signal in order to push the signal to either the positive or negative side. Clamper circuit is a combination of a resistor along with a diode and capacitor. It sometimes also employs dc battery so as to have an additional shift in the signal level.

Clamper circuits are constructed in a similar manner as that of clipper circuits. However, clamper includes an extra charging element that is the capacitor in its circuitry. The combination of resistor and capacitor in the clamper circuit is used to maintain different dc level at the output of the clamper.

Operating principle of Clamper circuits

As we have already discussed a clamper consist of the capacitor and a diode in shunt connection with the load.

The working of clamper circuits depends on the variation in the time constant of the capacitor. This variation is the outcome of changing the current path of the diode with the change in input signal polarity.

Here, the magnitude of the time constant is

$$\tau = RC$$

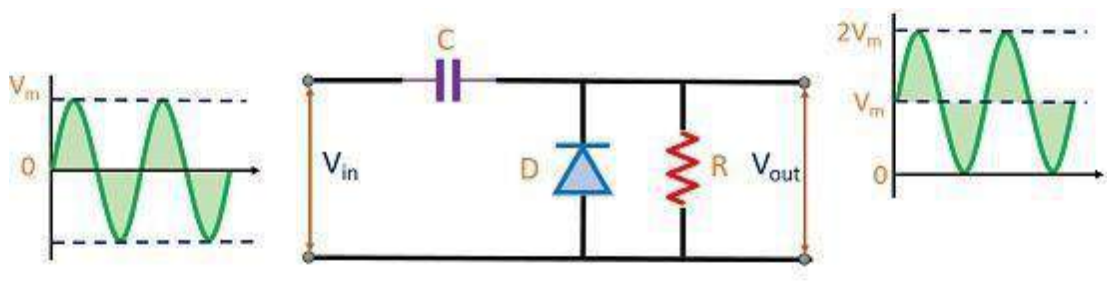
this is chosen large enough in order to assure that voltage across the capacitor does not discharge consequently at the non-conducting interval of the diode. But, such a discharge takes place only when the load resistance is very large. This permits the capacitor to take larger discharge time. Conversely, a smaller value of the capacitor is chosen so that it will charge rapidly at the time of conduction of the diode.

Classification of Clamper Circuits-

Clampers are classified in the following groups:

<u>Positive Clamper circuit</u>

The figure below shows the circuit of a positive clamper-



Positive Clamper circuit

As we can see here, the diode is in parallel connection with the load. So we can say reverse biasing of the diode will provide the output at the load.

Initially, the positive half of the applied input signal reverse biases the diode but the capacitor is not still charged. So, at this period of time output will not be considered.

For, the negative half of the AC signal, the capacitor now gets fully charged up to the peak of the AC signal but with inverse polarity. This negative half forward biases the diode that results in the flow of the forward current through the diode. The next positive half then reverse biases the diode due to which signal will appear at the output.

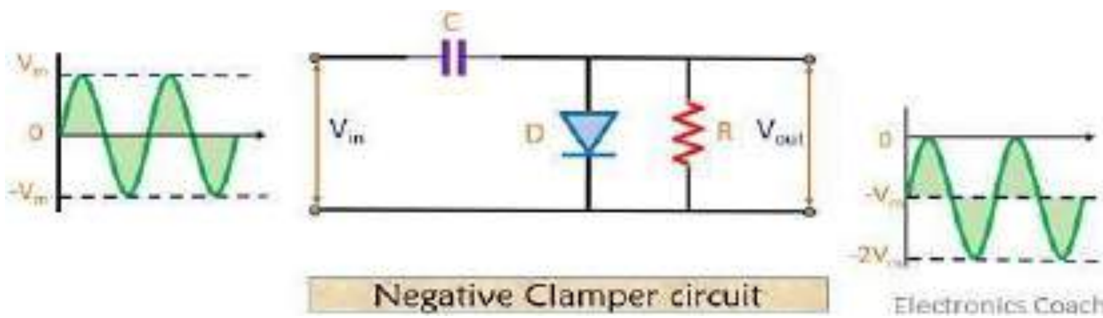
At the beginning of the positive half of the AC signal, the diode is in the non-conducting state that results in discharging of capacitor charge. So, at the output, we will have the summation of the voltage stored across the capacitor and applied the AC input signal. This is given by

$$V_o = V_m + V_m = 2V_m$$

Here as we can in the output waveform shown above, the signal level is shifted upward or positive side. Hence it is named as positive clamper.

Negative Clamper circuit

Let's have a look at the figure shown below of negative clamper in order to understand the detailed operation-



At the time when positive half of the AC input is applied, the diode comes to forward bias condition that results in no-load current at the output. However, a forward current flows through the diode that charges the capacitor to the peak of the ac signal but again with inverse polarity. The capacitor here is charged up to the forward biased condition of the diode.

When negative half of the AC signal is applied, the diode now becomes reverse biased. This allows load current to appear at the output of the circuit. Now, this non-conducting state of the diode discharges the capacitor. So, at the output, a summation of capacitor voltage along with the input voltage is achieved.

Hence at the output, we have,

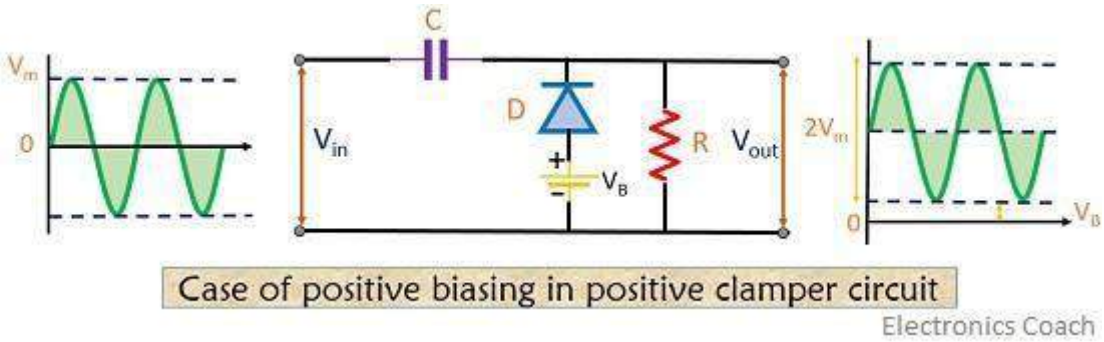
$V_o = -V_m - V_m = -2V_m$

This results in the downward shift of the signal. Therefore, it is termed as negative clamper circuit.

Positive clamper circuit with biasing

This is basically done to introduce an additional shift in the level of the signal. Here, the biasing provided to the circuit is of two types. It can be a positively or negatively biased circuit. So, we will discuss both the cases separately.

1. Case of positive biasing



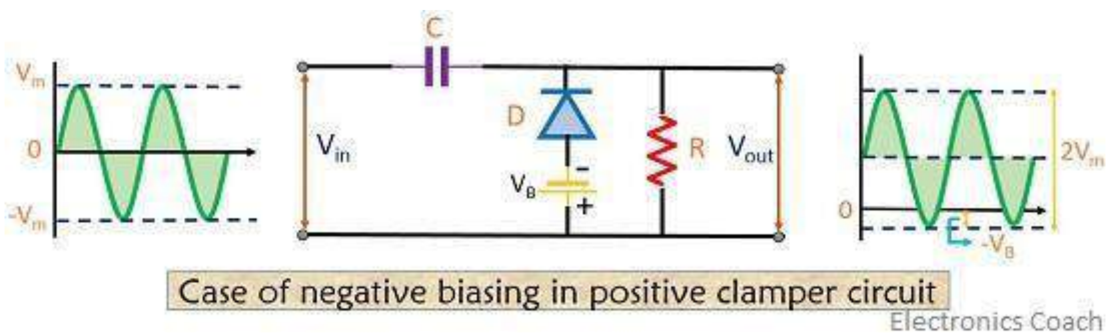
Case of positive biasing in positive clamper circuit

The working is almost similar to the positive unbiased case but here an additional voltage is provided so as to have an additional shift in the level of the signal.

When positive half of the input signal is applied, the diode is reverse biased due to ac input but is forward biased due to battery voltage. So, until the voltage of a battery is greater than the ac input the diode conducts. This forward current through the diode charges the capacitor but with the battery voltage. As the ac input surpasses battery voltage, the diode now gets reverse biased and hence conduction through the diode stops.

On the application of the negative half of the input signal, the diode is now forward biased due to both ac input and battery voltage and starts conducting. This charges the capacitor with voltage summation of ac input along with battery voltage. Hence such an output voltage level is achieved.

2. Case of negative biasing



Case of negative biasing in positive clamper circuit

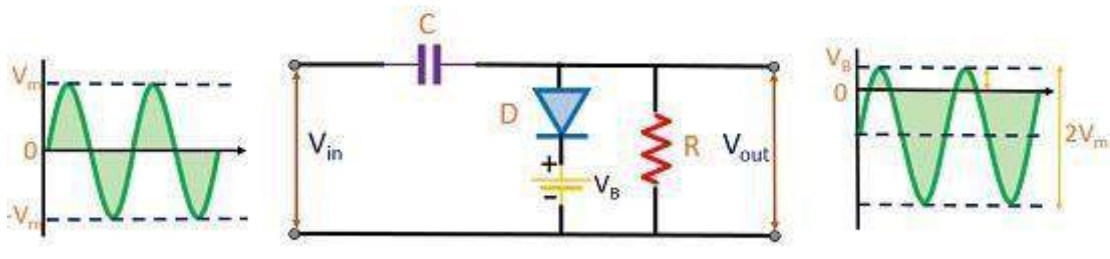
At the time of positive half of the AC signal, the diode gets reverse biased by both ac input and battery voltage. Due to this current flows through the load and combinedly maintain the voltage level.

At the time of the negative half, the diode is in the forward biased condition due to ac input but is in reverse biased condition due to battery voltage. So, the diode conducts only when the AC input dominates the battery voltage. This charges the capacitor hence we get a shifted signal at the output.

Negative Clamper circuit with biasing

In a similar way as previous, positive and negative biasing is provided to the negative clamper circuit. Let's now move further and discuss both the case separately.

1. Case of positive biasing



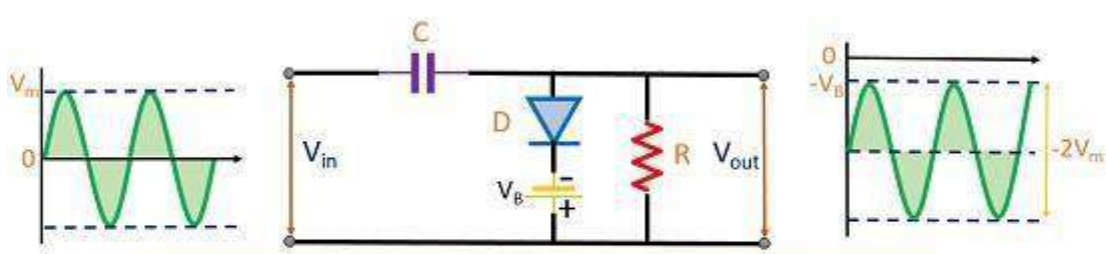
Case of positive biasing in negative clamper circuit

Electronics Coach

As we have already discussed that a negative clamper shifts the signal downward. However, in the case of the positively biased negative clamper, the signal is somewhat raised to a positive level due to the positively applied battery voltage. When positive half of the AC signal is applied, the diode is in the forward biased state due to ac supply but is reverse biased because of battery voltage. So, the diode conducts when ac supply surpasses battery voltage.

Moving further during the negative half, the diode is now in reverse biased state by cause of both the AC supply and battery voltage. This non-conducting state of the diode discharges the capacitor. Thus, the voltage across the capacitor appears at the output.

2. Case of negative biasing



Case of negative biasing in negative clamper circuit

Electronics Coach

At the time of positive half of ac input, the diode gets forward biased by the cause of ac input and battery voltage. This starts conduction through the diode. Resultantly charges the capacitor.

At the time of the negative half, the diode gets reverse biased but will still conduct due to forward biased condition applied by the battery. The diode current flows until the battery voltage is more than the ac input supply. The time when ac input surpasses battery voltage, the diode gets reverse biased and the capacitor discharges. Thus the voltage across capacitor appears at the load.

Applications of Clamper Circuits

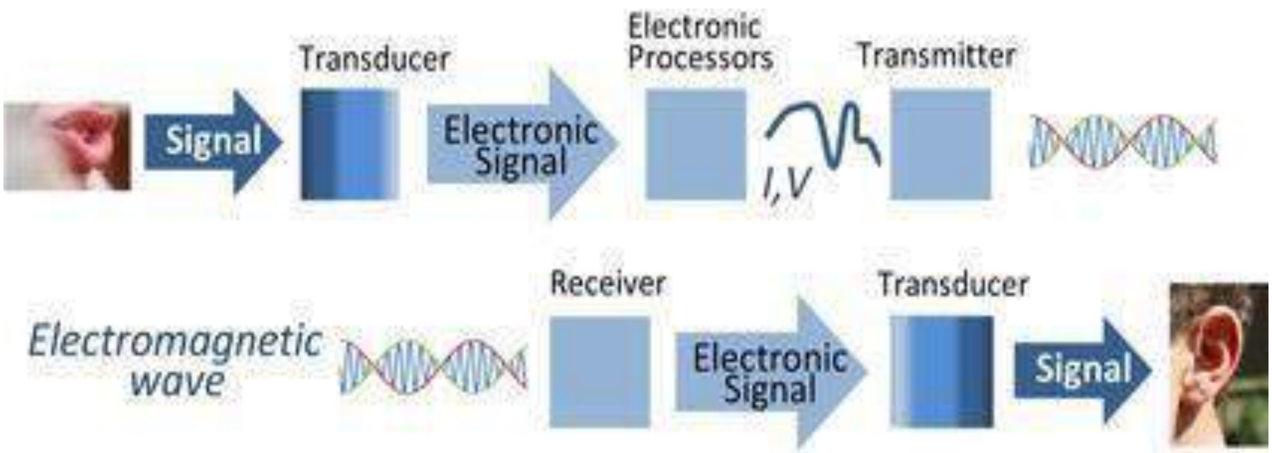
1. Clampers are used to identify the polarity of the circuits.
2. These circuits are used as voltage doublers and help in eliminating distortions.
3. Reverse recovery time can be improved using Clampers.
4. Comparison Chart

PARAMETER	CLIPPER	CLAMPER
Definition	Clipper delimit the amplitude of the output voltage.	Clamper shifts the DC level of the output voltage.
Output Voltage	Less than the input voltage.	Multiples of input voltage.
Energy storage component	Not required	Requires (Capacitor is used as energy storage element)
Shape of Output Waveform	Shape changes (Rectangular, sinusoidal, triangular etc.)	Shape remains same as input waveform.
DC Level	Remains same	DC level get shifted
Applications	In transmitters, receivers, amplitude selector, noise limiter etc.	In voltage multiplying circuits, Sonar, Radar system etc.

UNIT- 2: SPECIAL SEMICONDUCTOR DEVICES

**TRANSD
UCER:-**

- ✓ Transducer is a device which converts one form of energy into another form i.e.; the given non-electrical energy is converted into an electrical energy.
- ✓ Common examples include microphones, loudspeakers, thermometers, position and pressure sensors and antenna.



- ✓ Efficiency is an important consideration in any transducer.
  Transducer efficiency is defined as the ratio of the power output in the desired form to the total power input.
- ✓ Mathematically,
  If P represents the total power input and Q represents the power output in the desired form.

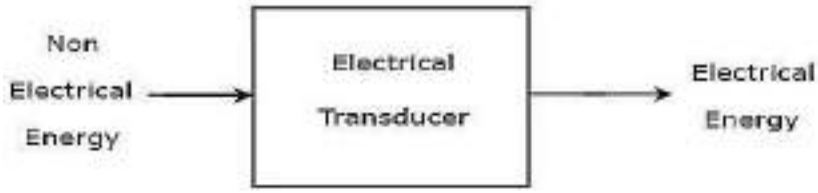
$$E = \frac{Q}{P}$$

- ✓ In percentage,

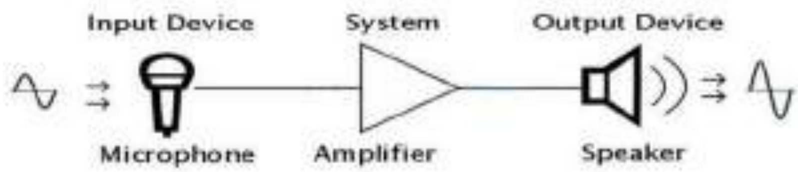
$$E\ (\%) = \frac{Q}{P}*100$$

- ✓ No transducer is 100% efficient; some power is always lost in the conversion process.
  Usually this loss is dissipated in the form of heat.
- ✓ An antenna is also a transducer which converts electrical signal into electromagnetic waves and vice-versa.

- ✓ The transducer, which converts non-electrical form of energy into electrical form of energy, is known as electrical transducer.
- ✓ The block diagram of electrical transducer is shown in below figure.



- ✓ As shown in the figure, electrical transducer will produce an output, which has electrical energy. The output of electrical transducer is equivalent to the input, which has non-electrical energy.



DIFFERENT TYPES OF TRANSDUCERS:-

- ✓ First let's discuss about two main types of transducers which we use every day in our industrial life.
- ✓ They are:-

  a) Active transducers
  b) Passive transducers

a) Active transducers

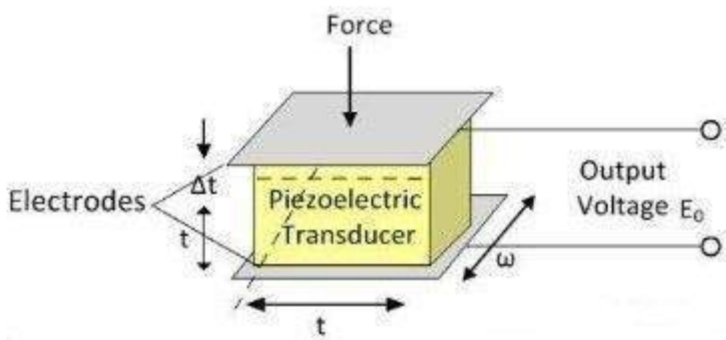
- ✓ Active transducers are those which convert one form of energy into another form (electrical) without requiring any external source of power.
- ✓ These transducers draw the energy needed for their operation from the measuring system itself.
- ✓ Following are the examples of active transducers:-
- Piezo-electric crystals
  This converts charges generated by application of force into electric potential.
- Tachogenerator
  These are basically used to measure angular velocity
  - Thermocouple

Temperature measurement is accomplished using thermocouples.

- Photovoltaic cell

  It converts light into electrical energy.

✓ The block diagram of active transducer is shown in below figure.



b) Passive transducers

✓ Passive transducers are those transducers which convert a form of energy into another (electrical) by making use of an external source of power.

✓ This transducer induces variation in the parameters associated with the electrical circuits, with the variation in the applied input signal.

✓ Following are the examples of passive transducers.

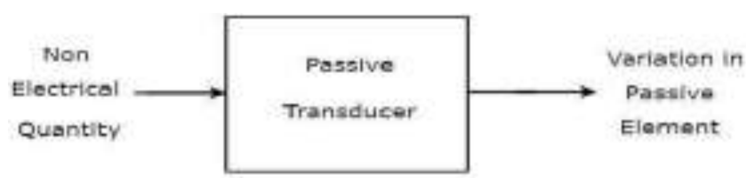
- Potentiometer

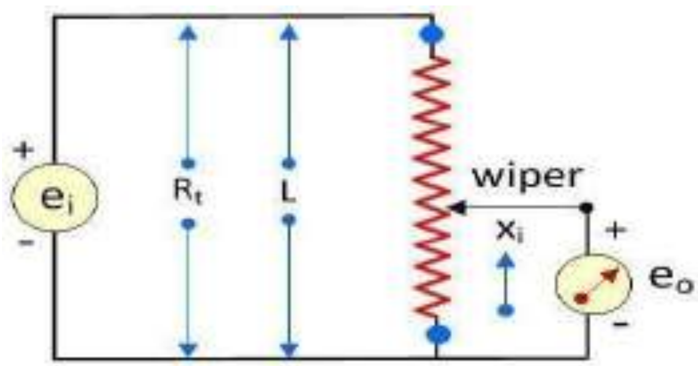
  It is a device that converts displacement into voltage.

- Thermistor

  These produce voltage with change in temperature

✓ The block diagram of passive transducer is shown in below figure.

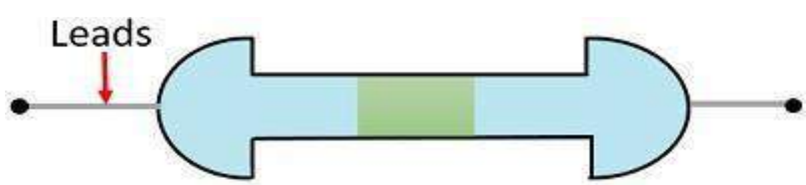




Linear Potentiometer (Pot), a passive transducer

THERMISTOR:-

- ✓ Thermistor, a semiconductor device is a type of passive transducer in which variation in temperature causes a corresponding change in resistance.
- ✓ Thus, variation in temperature produces an analogue voltage. As these are thermally sensitive resistors thus also termed as thermal resistors.
- ✓ The figure below shows the rod form of a Thermistor

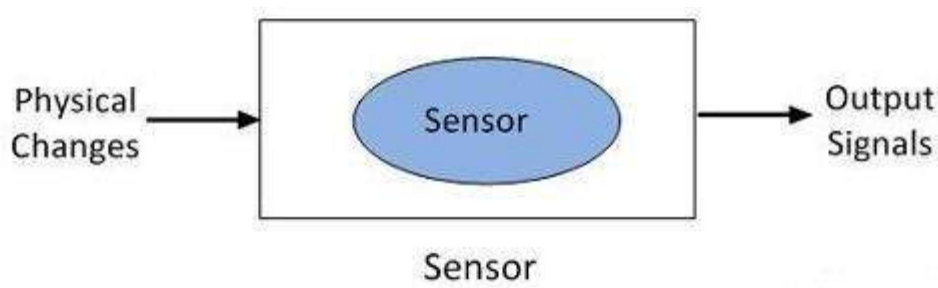


Rod form of thermistor

- ✓ Due to their temperature sensitive nature, thermistors have various applications in temperature measurement field.
- ✓ It can be a positive temperature coefficient thermistor or negative temperature coefficient thermistor depending on the variation of resistance with respect to temperature.
- ✓ An extremely non-linear characteristic is exhibited by the thermistor for resistance versus temperature curve. As it is inexpensive and highly sensitive device thus has numerous applications.

SENSOR:

- ✓ A sensor is a physical device that senses a physical quantity and then converts it into signals which can be read by an instrument or the user.
- ✓ The specific input could be light, heat, motion, moisture, pressure etc.
- ✓ The output is generally a signal that is converted to human-readable display at the sensor location or transmitted electronically over a network for reading or further processing.
- ✓ Sensors are sophisticated devices that are frequently used to detect and respond to electrical or optical signals.
- ✓ A Sensor converts the physical parameter (for example: temperature, blood pressure, humidity, speed, etc.) into a signal which can be measured electrically.
- ✓ Both the input and output quantities of a Sensor are Physical i.e. non-electrical in nature.



- ✓ The sensors have many applications in the electronics equipment. The few of them are explained below:-

- The motion sensors are used in the home security system and the automation door system.
- The photo sensor senses the infrared or ultraviolet light.
- The accelerometer sensor used in mobile for detecting the screen rotations.

CLASSIFICATION OF SENSORS:-

Based on the applications of sensors, their classification can be made as follows.

I. Displacement, Position and Proximity Sensors

1. Resistive Element or Potentiometer
2. Capacitive Elements
3. Strain Gauged Element
4. Inductive Proximity Sensors
5. Eddy Current Proximity Sensors
6. Differential Transformers
7. Optical Encoders
8. Hall Effect Sensors
9. Pneumatic Sensors
10. Proximity Switches
11. Rotary Encoders

II. Temperature Sensors

1. Thermistors
2. Thermocouple
3. Bimetallic Strips
4. Resistance Temperature Detectors
5. Thermostat

III. Light Sensors

1. Photo Diode
2. Phototransistor
3. Light Dependent Resistor

IV. Velocity and Motion

1. Pyroelectric Sensors
2. Tachogenerator
3. Incremental encoder

V. Fluid Pressure

1. Diaphragm Pressure Gauge
2. Tactile Sensor
3. Piezoelectric Sensors
4. Capsules, Bellows, Pressure Tubes

VI. Liquid Flow and Level

1. Turbine Meter
2. Orifice Plate and Venturi Tube

VII. IR Sensor

1. Infrared Transmitter and Receiver Pair

VIII. Force

1. Strain Gauge
2. Load Cell

IX. Touch Sensors

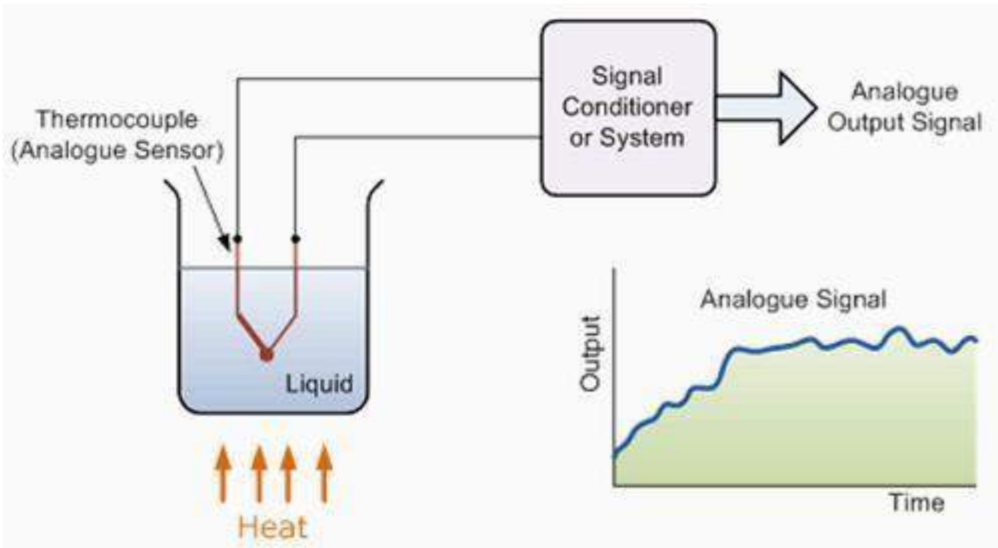
1. Resistive Touch Sensor
2. Capacitive Touch Sensors

X. UV Sensors

1. Ultraviolet Light Detector
2. Photo Stability Sensors
3. UV Photo Tubes
4. Germicidal UV Detectors

ANALOGUE SENSORS

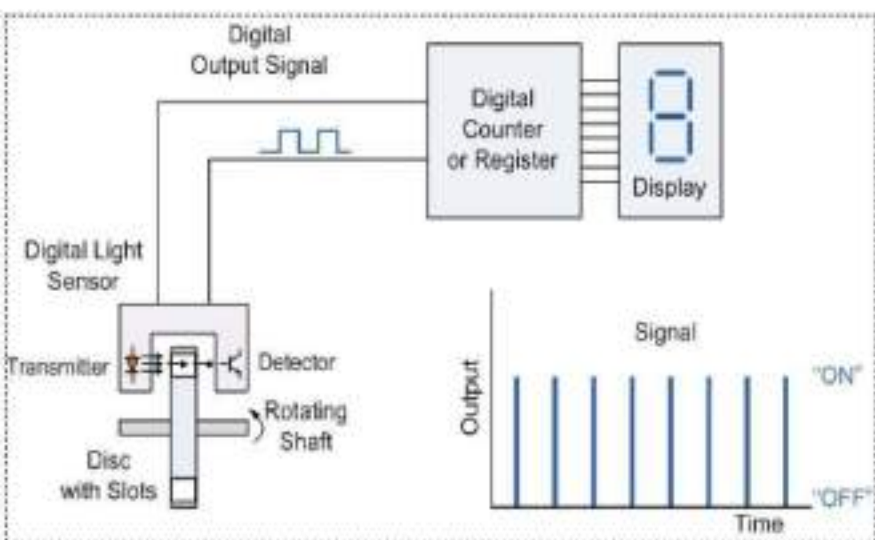
- ✓ An analogue sensor produces continuously varying output signals over a range of values.
- ✓ Usually the output signal is voltage and this output signal is proportional to the measurand.
- ✓ The quantity that is being measured like speed, temperature, pressure, strain, etc. are all continuous in nature and hence they are analogue quantities.



✓ A thermocouple or a thermometer is an analog sensor. The following setup is used to measure the temperature of the liquid in the container using a thermocouple.

DIGITAL SENSORS

✓ A digital sensor produces discrete digital signals.
✓ The output of a digital sensor has only two states, namely 'ON' and 'OFF'. ON is logic 1 and OFF is logic 0.
✓ A push button switch is the best example of a digital sensor. In this case, the switch has only two possible states: either it is ON when pushed or it is OFF when released or not pushed.



✓ The following setup uses a light sensor to measure the speed and produces a digital signal.
✓ In general, the accuracy of a digital sensor is high when compared to an analogue sensor.

- ✓ The accuracy depends on the number of bits that are used to represent the measurand.
- ✓ Higher the number of bits, the greater is the accuracy.

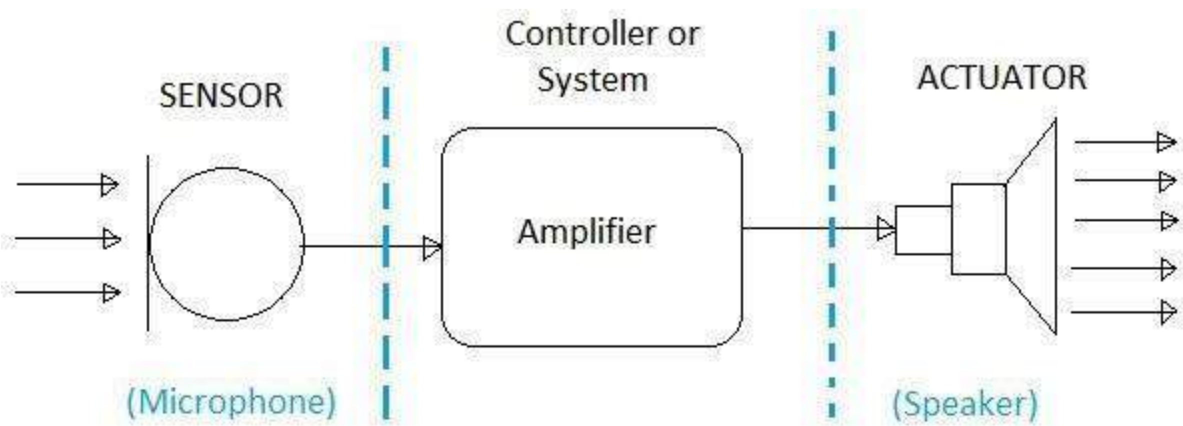
- ✓ Here are a few examples of the many different types of sensors:

a) In a mercury-based glass thermometer, the input is temperature. The liquid contained expands and contracts in response, causing the level to be higher or lower on the marked gauge, which is human-readable.

b) Motion sensors in various systems including home security lights, automatic doors and bathroom fixtures typically send out some type of energy, such as microwaves, ultrasonic waves or light beams and detect when the flow of energy is interrupted by something entering its path.

c) A photo sensor detects the presence of visible light, infrared transmission (IR), and/or ultraviolet (UV) energy.

- ✓ Both Microphone and Loudspeaker are Transducers in the sense that a microphone converts sound energy into electrical energy and a loud speaker converts electrical energy into sound energy.



BARRETTER:

- A bolometer element with a positive temperature coefficient of resistance, used to detect and measure power at radio, microwave, infrared, and optical frequencies. The temperat ure of the barretter increases when electromagnetic energy is absorbed. Barretters are m ade of metal; therefore, the electrical resistance increases when the temperature increas es. The resulting resistance change of the barretter is measured by using direct-current or low-frequency instruments.

- The barretter resistance is selected to absorb most of the power when the barretter is m ounted as a termination in a waveguide or coaxial transmission line. A barretter can be m ade to detect power at optical and infrared frequencies by using a very thin metal ribbon blackened to absorb light.
- Barretters with less sensitivity and accuracy for use at radiofrequencies can be made by using low current fuses made with fine wires.
- A meter can be made to measure high frequency signal amplitudes using a barretter. The temperature and hence the resistance of a barretter can change at audio frequency rates. But the time constant of a barretter is too great for the resistance to vary at radio frequency rates.
- A radio or microwave frequency current modulated at a low frequency will cause the barretter resistance to follow the low frequency signal. If a direct current voltage is applied to the barretter while the modulated radio frequency current is also applied, the varying resistance will produce a current which follows the modulation.
- The low frequency current can be coupled to the input of an audio amplifier tuned to the modulation frequency by using an audio transformer. The output of the audio amplifier may be rectified to drive a direct current meter. The meter then indicates the relative amplitude of the radio frequency or microwave signal

ZENER DIODE:-

Zener diode is basically like an ordinary PN junction diode but normally operated in reverse biased condition. But ordinary PN junction diode connected in reverse biased condition is not used as Zener diode practically. A Zener diode is a specially designed, highly doped PN junction diode.

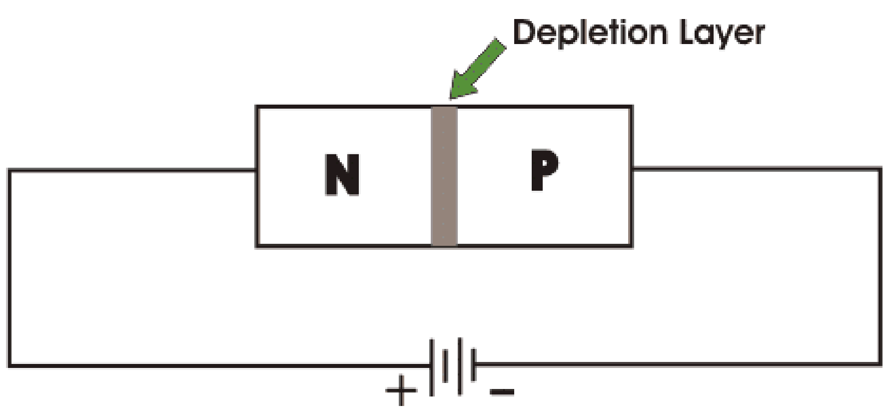
Working Principle of Zener Diode

When a PN junction diode is reverse biased, the depletion layer becomes wider. If this reverse biased voltage across the diode is increased continually, the depletion layer becomes more and more wide. At the same time, there will be a constant reverse saturation current due to minority carriers.

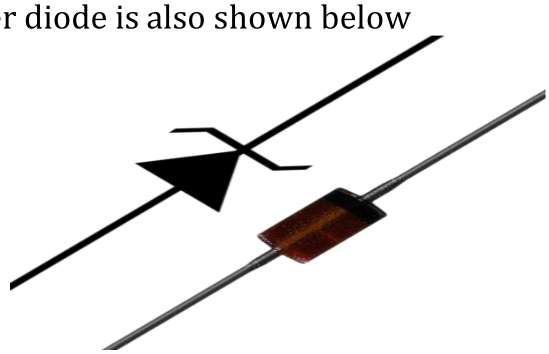
After certain reverse voltage across the junction, the minority carriers get sufficient kinetic energy due to the strong electric field. Free electrons with sufficient kinetic energy collide with stationary ions of the depletion layer and knock out more free electrons. These newly created free electrons also get sufficient kinetic energy due to the same electric field, and they create more free electrons by collision cumulatively. Due to this commutative phenomenon, very soon, huge free electrons get created in the depletion layer, and the entire diode will become conductive. This type of breakdown of the depletion layer is known as avalanche breakdown, but this breakdown is not quite sharp. There is another type of breakdown in depletion layer which is sharper compared to avalanche breakdown, and this is called Zener breakdown. When a PN junction is diode is highly doped, the concentration of impurity atoms will be high in the crystal. This higher concentration of impurity atoms causes the higher concentration of ions in the depletion layer hence for same applied reverse biased voltage, the width of the depletion layer becomes thinner than that in a normally doped diode.

Due to this thinner depletion layer, voltage gradient or electric field strength across the depletion layer is quite high. If the reverse voltage is continued to increase, after a certain applied voltage, the electrons from the covalent bonds within the depletion region come out and make the depletion region conductive. This breakdown is called Zener breakdown. The voltage at which this breakdown occurs is called Zener voltage. If the applied reverse voltage across the diode is more than Zener voltage, the diode provides a conductive path to the current through it hence; there is no chance of further avalanche breakdown in it. Theoretically, Zener breakdown occurs at a lower voltage level then avalanche breakdown in a diode, especially doped for Zener breakdown. The Zener breakdown is much sharper than avalanche breakdown. The Zener voltage of the diode gets adjusted during manufacturing with the help of required and proper doping. When a Zener diode is connected across a voltage source, and the source voltage is more than Zener voltage, the voltage across a Zener diode remain fixed irrespective of the source voltage. Although at that condition current through the diode can be of any value depending on the load connected with the diode. That is why we use a Zener diode mainly for controlling voltage in different circuits.

A diode connected in reverse bias position in a circuit is shown below,

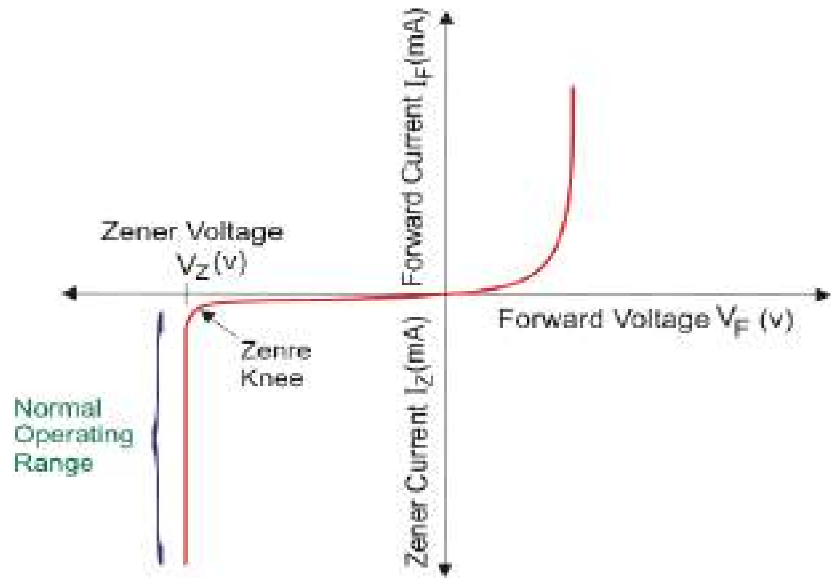


The circuit symbol of a Zener diode is also shown below



Characteristics of a Zener Diode

The V-I characteristics of a Zener diode is given below:



The above diagram shows the V-I characteristics of a Zener diode. When the diode is connected in forward bias, this diode acts as a normal diode but when the reverse bias voltage is greater than Zener voltage, a sharp breakdown takes place. In the V-I characteristics above V_z is the Zener voltage. It is also the knee voltage because at this point the current increases very rapidly.

TUNNEL DIODE:-

A tunnel diode (also known as a Esaki diode) is a type of semiconductor diode that has effectively "negative resistance" due to the quantum mechanical effect called tunneling. Tunnel diodes have a heavily doped pn junction that is about 10 nm wide. The heavy doping results in a broken band gap, where conduction band electron states on the N-side are more or less aligned with valence band hole states on the P-side.

The application of transistors in a very high in frequency range is hampered due to the transit time and other effects. Many devices use the negative conductance property of semiconductors for these high frequency applications. A tunnel diode is one of the most commonly used negative conductance devices. It is also known as Esaki diode after L. Esaki for his work on this effect.

The concentration of dopants in both p and n region is very high, at around 10^{24} – 10^{25} m^{-3}. The pn junction is also abrupt. For this reasons, the depletion layer width is very small. In the current voltage characteristics of tunnel diode, we can find a negative slope region when a forward bias is applied.

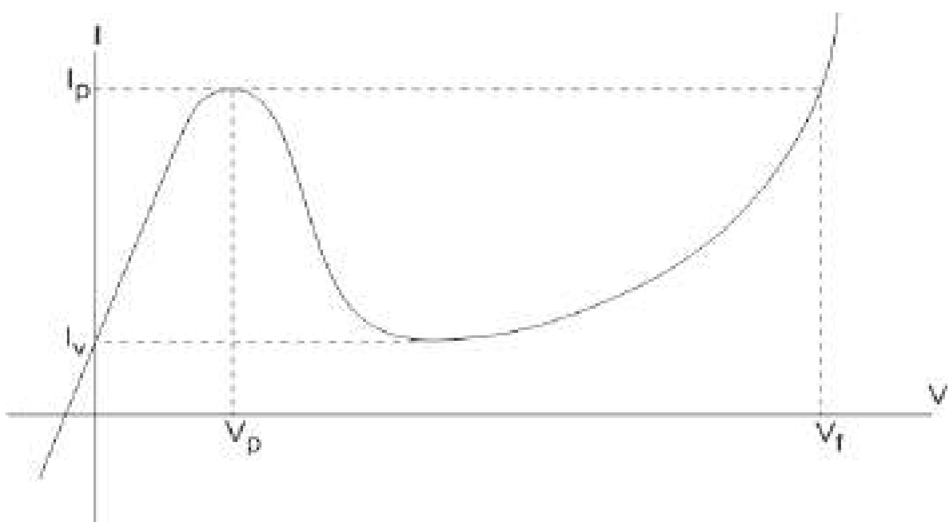
The name “tunnel diode” is due to the quantum mechanical tunneling is responsible for the phenomenon that occurs within the diode. The doping is very high so at absolute zero temperature the Fermi levels lies within the bias of the semiconductors.

Characteristics of Tunnel Diode

When reverse bias is applied the Fermi level of the p-side becomes higher than the Fermi level of n-side. Hence, the tunneling of electrons from the balance band of p-side to the conduction band of n-side takes place. With the interments of the reverse bias the tunnel current also increases.

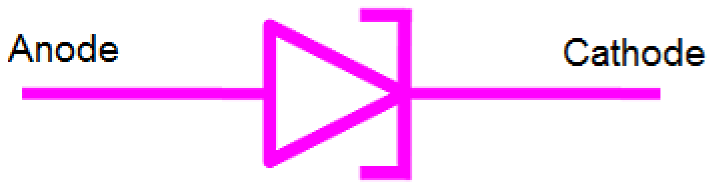
When forward bias is applied the Fermi level of n-side becomes higher that the Fermi level of p-side, thus the tunneling of electrons from the n-side to p-side takes place. The amount of the tunnel current is very large than the normal junction current. When the forward bias is increased, the tunnel current is increased up to certain limit.

When the band edge of n-side is the same as the Fermi level in p-side, the tunnel current is maximum with the further increment in the forward bias the tunnel current decreases and we get the desired negative conduction region. When the forward bias is raised further, normal pn junction current is obtained which is exponentially proportional to the applied voltage. The V-I characteristics of the tunnel diode is given,
The negative resistance is used to achieve oscillation.



Tunnel Diode Symbol

The symbol for a tunnel diode is shown below.



Tunnel diode symbol

Tunnel Diode Applications

Tunnel diode is a type of sc diode which is capable of very fast and in microwave frequency range. It was the quantum mechanical effect which is known as tunneling. It is ideal for fast oscillators and receivers for its negative slope characteristics. But it cannot be used in large integrated circuits – that's why it's an applications are limited.

When the voltage is first applied current starts flowing through it. The current increases with the increase of voltage. Once the voltage rises high enough suddenly the current again starts increasing and tunnel diode stars behaving like a normal diode. Because of this unusual behavior, it can be used in number of special applications started below.

Oscillator Circuits:

Tunnel diodes can be used as high frequency oscillators as the transition between the high electrical conductivity is very rapid. They can be used to create oscillation as high as 5Gz. Even they are capable of creativity oscillation up to 100 GHz in a appropriate digital circuits.

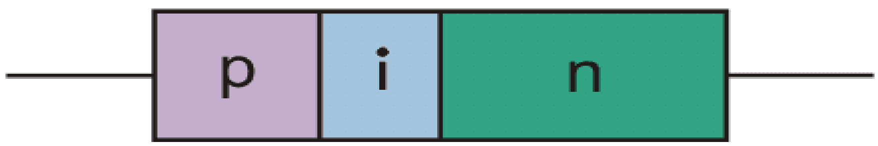
Used in Microwave Circuits:

Normal diode transistors do not perform well in microwave operation. So, for microwave generators and amplifiers tunnel diode are used. In microwave waves and satellite communication equipment they were used widely, but lately their usage is decreasing rapidly, as transistors which operate in this frequency range are becoming available.

Tunnel diodes are resistant to the effects of magnetic fields, high temperature and radioactivity. That's why these can be used in modern military equipment. These are used in nuclear magnetic resource machine also. But the most important field of its use satellite communication equipment.

PIN DIODE:

A Pin diode is a special type of diode that contains an undoped intrinsic semiconductor between the p-type semiconductor and n-type semiconductor regions. It differs from a normal diode in the sense that it has an extra layer in between the p and the n junctions. By an intrinsic layer, we mean a pure crystal of silicon or germanium without any doping in it. This layer does not conduct electric current well. The p-type and n-type layer is heavily doped as they are used for ohmic contacts.



PIN Diode

It's like having a p region, followed by an intrinsic region and then an N region, thus making it a PIN diode and hence derived the name from the same.

The symbol of pin diode is as follows:

Symbol of PIN Diode

Construction of a PIN Diode

A pin diode has an intrinsic undoped layer (having high resistivity) sandwiched between a PN junction, we will now look at the construction of the diode in detail. Pin diode can be constructed in two different structures: Mesa structure and planar structure. In mesa structure, layers which are already doped are grown onto the substrate (intrinsic layer). The amount of doping and thickness of the layer can be controlled as per the requirement. In a planar structure, an epitaxial layer is grown onto the substrate material and the p+ region is created either by ion implantation or diffusion.

Working of the PIN Diode

Although the working of a pin diode is similar to a regular diode, there are some differences due to the additional intrinsic layer. Pin diode is an inferior rectifier but is widely suitable for switches, attenuators etc.

Forward biased operation of PIN Diode

When the PIN diode is subjected to forward bias, the depletion region at the p-n junction reduces. With the reduction of the depletion region, current starts flowing through the diode. The PIN diode acts as a variable resistance when operated in forward bias. A high electric field is developed across the junction and this speeds up the transport of charge carriers from the P region to the N region. This helps in faster operation and therefore pin diode is used for high-frequency applications.

The width of the depletion region increases, when the pin diode operates in reverse biased condition. At a certain reverse bias voltage, the entire intrinsic layer will be swept out of charge carriers. This voltage is called the swept in voltage. The value is -2v. It is used for switching purposes while in reverse bias.

Characteristics of PIN Diode

At a lower level of reverse bias, the depletion layer becomes fully depleted. The capacitance of the pin diode becomes independent of the level of bias once the depletion layer is fully depleted. This is because there is very little net charge in the intrinsic layer. The leakage of RF signal is lower than other diodes because the level of capacitance is typically lower.

In forward bias, the diode behaves as a resistor than a non-linear device and produces no rectification or distortion. The value of the resistance depends on the bias voltage. Pin diode is used as RF switch or variable resistor as they produce fewer distortions than a normal diode.

Application of PIN Diode

Pin diodes are used for a number of applications. Some of the important usages of these types of diodes are:

- RF switch: As discussed above, a pin diode can be used as an RF switch. Due to the intrinsic layer between the PN junctions, the level of capacitance decreases. As a consequence, the level of isolation is increased when the diode is reversed bias. This makes it an ideal RF switch.
- High Voltage Rectifier: Due to the intrinsic layer, the distance between the p-n junctions is increased and thus allowing higher voltage tolerance between the junctions.
- Photo detector: The conversion of light into electric current depends on the depletion region of the photodiode. In a pin diode, the depletion region is increased by adding an intrinsic layer between the p-n junctions. Due to the increase in the depletion region, the volume of conversion is increased and the efficiency of the photodiode too. Hence pin diode is used as photodiodes.

UNIT-3: RECTIFIER CIRCUITS & FILTERS

RECTIFIERS AND CLASSIFICATION OF RECTIFIER:-

Rectifier is an electronic device which converts the alternating current to unidirectional current, in other words rectifier converts the AC voltage to DC voltage. We use rectifier in almost all the electronic devices mostly in the power supply section to convert the main voltage into DC voltage. Every electronic device will work on the DC voltage supply only.

Rectifiers are classified according to the period of conduction.

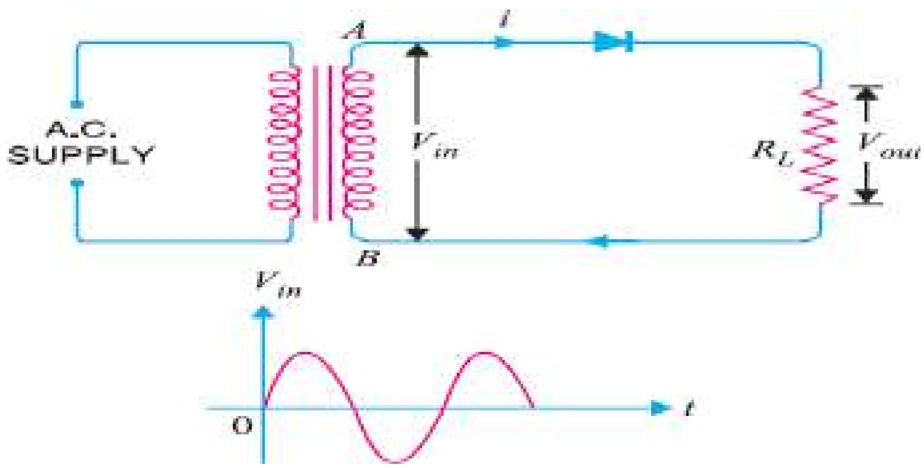
They are

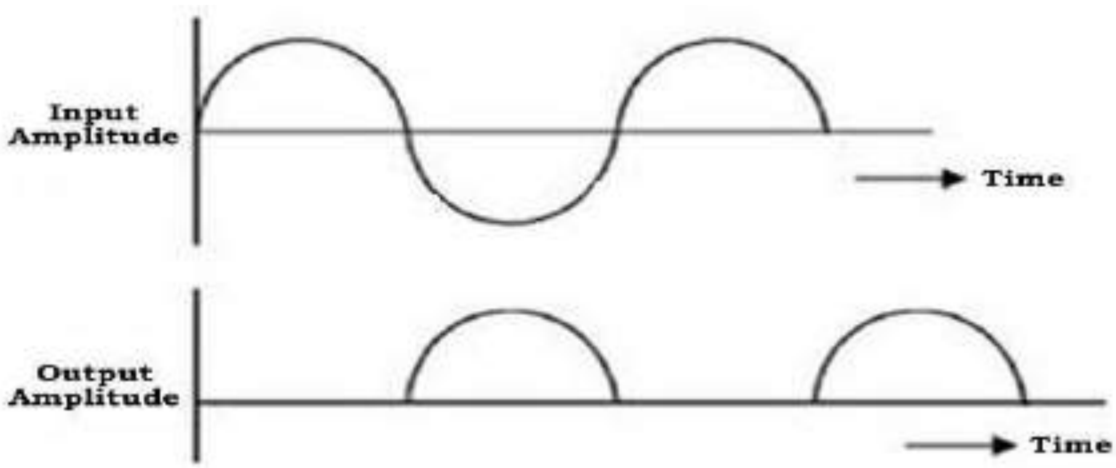
1. Half Wave Rectifier
2. Full Wave Rectifier

ANALYSIS OF HALF WAVE, FULL WAVE CENTRE TAPPED AND BRIDGE RECTIFIERS:

Half Wave Rectifier:

The half wave rectifier is a type of rectifier that rectifies only half cycle of the waveform. This article describes the half wave rectifier circuit working. The half rectifier consist a step down transformer, a diode connected to the transformer and a load resistance connected to the cathode end of the diode. The circuit diagram of half wave rectifier is shown below:





The main supply voltage is given to the transformer which will increase or decrease the voltage and give to the diode. In most of the cases we will decrease the supply voltage by using the step down transformer here also the output of the step down transformer will be in AC. This decreased AC voltage is given to the diode which is connected serial to the secondary winding of the transformer, diode is electronic component which will allow only the forward bias current and will not allow the reverse bias current. From the diode we will get the pulsating DC and give to the load resistance RL.

Working of Half Wave Rectifier:

The input given to the rectifier will have both positive and negative cycles. The half rectifier will allow only the positive half cycles and omit the negative half cycles. So first we will see how half wave rectifier works in the positive half cycles.

Positive Half Cycle:

- In the positive half cycles when the input AC power is given to the primary winding of the step down transformer, we will get the decreased voltage at the secondary winding which is given to the diode.
- The diode will allow current flowing in clock wise direction from anode to cathode in the forward bias (diode conduction will take place in forward bias) which will generate only the positive half cycle of the AC.
- The diode will eliminate the variations in the supply and give the pulsating DC voltage to the load resistance RL. We can get the pulsating DC at the Load resistance.

Negative Half Cycle:

- In the negative half cycle the current will flow in the anti-clockwise direction and the diode will go in to the reverse bias. In the reverse bias the diode will not conduct so, no current in flown from anode to cathode, and we cannot get any power at the load resistance.
- Only small amount of reverse current is flown from the diode but this current is almost negligible. And voltage across the load resistance is also zero.

Characteristics of Half Wave Rectifier:

There are some characteristics to the half wave rectifier they are

1. Efficiency: The efficiency is defined as the ratio of input AC to the output DC.

Efficiency, $\eta = P_{dc} / P_{ac}$

DC power delivered to the load, $P_{dc} = I^2_{dc} R_L = (I_{max/pi})^2 R_L$

AC power input to the transformer, P_{ac} = Power dissipated in junction of diode + Power dissipated in load resistance R_L

$= I^2_{rms} R_F + I^2_{rms} R_L = \{I^2_{MAX}/4\}[R_F + R_L]$

Rectification Efficiency, $\eta = P_{dc} / P_{ac} = \{4/ ^2\}[RL/ (R_F + R_L)] = 0.406/\{1+ R_{F/RL} \}$

If R_F is neglected, the efficiency of half wave rectifier is 40.6%.

2. Ripple factor: It is defined as the amount of AC content in the output DC. It is nothing but amount of AC noise in the output DC. Less the ripple factor, performance of the rectifier is more. The ripple factor of half wave rectifier is about 1.21. It can be calculated as follows:

The effective value of the load current I is given as sum of the rms values of harmonic currents I1, I2, I3, I4 and DC current Idc.

$$I^2 = I^2_{dc}+I^2_1+I^2_2+I^2_4 = I^2_{dc} + I^2_{ac}$$

Ripple factor, is given as $\gamma = I_{ac} / I_{dc} = (I^2 - I^2_{dc}) / I_{dc} = \{(I_{rms} / I_{dc}^2)-1\} = K_f^2 - 1)$

Where K_f is the form factor of the input voltage. Form factor is given as

$$K_f = I_{rms} / I_{avg} = (I_{max}/2) / (I_{max}/pi) = pi/2 = 1.57$$

So, ripple factor, $\gamma = (1.57^2 - 1) = 1.21$

3. Peak Inverse Voltage: It is defined as the maximum voltage that a diode can with stand in reverse bias. During the reverse bias as the diode do not conduct total voltage drops across the diode. Thus peak inverse voltage is equal to the input voltage Vs.

4. Transformer Utilization Factor (TUF): The TUF is defined as the ratio of DC power is delivered to the load and the AC rating of the transformer secondary. Half wave rectifier has around 0.287 and full wave rectifier has around 0.693.

✓ $TUF = \frac{P_{dc}}{P_{ac}}$

$$= \frac{(\frac{I_m}{\pi})^2 R_L}{\frac{V_m I_m}{2\sqrt{2}}}$$

$= 0.287$

5. Voltage regulation

The variation of d.c output voltage as function of d.c load current is called regulation.

$$V.R \text{ in \% age} = \frac{V_{NL} - V_{FL}}{V_{FL}} * 100$$

Where, V_{NL}=DC voltage across load resistance when minimum current flows through it.

V_{FL}=DC voltage across load resistance when maximum current flows through it.

6. Form factor

It is the ratio of the rms value to the average value.

$$\text{Form factor} = \frac{\text{Rms value}}{\text{Average value}} = \frac{Im/2}{Im/\pi} = 1.57$$

7. Output DC Voltage

The output voltage (V_{DC}) across the load resistor is denoted by:

$$V_{DC} = \frac{V_{Smax}}{\pi},\ where\ V_{Smax} = maximum\ amplitude\ of\ secondary\ voltage$$

8. RMS value of Half Wave Rectifier

To derive the RMS value of half wave rectifier, we need to calculate the current across the load. If the instantaneous load current is equal to $i_L = I_m Sin\omega t$, then the average of load current (I_{DC}) is equal to:

$$I_{dc} = \frac{1}{2\pi}\int_0^{\pi} I_m\, sin\omega t = \frac{I_m}{\pi}$$

Where I_m is equal to the peak instantaneous current across the load (I_{max}). Hence the output DC current (I_{DC}) obtained across the load is:

$$I_{DC} = \frac{I_{max}}{\pi},\ where\ I_{max} = maximum\ amplitude\ of\ dc\ current$$

For a half-wave rectifier, the RMS load current (I_{rms}) is equal to the average current (I_{DC}) multiple by $\pi/2$. Hence the RMS value of the load current (I_{rms}) for a half wave rectifier is:

$$I_{rms} = \frac{I_m}{2}$$

Where $I_m = I_{max}$ which is equal to the peak instantaneous current across the load.

Half wave rectifier is mainly used in the low power circuits. It has very low performance when it is compared with the other rectifiers.

FULL WAVE RECTIFIER:-

Full wave rectifier rectifies the full cycle in the waveform i.e. it rectifies both the positive and negative cycles in the waveform. This Full wave rectifier has an advantage over the half wave i.e. it has average output higher than that of half wave rectifier. The number of AC components in the output is less than that of the input.

The full wave rectifier can be further divided mainly into following types.

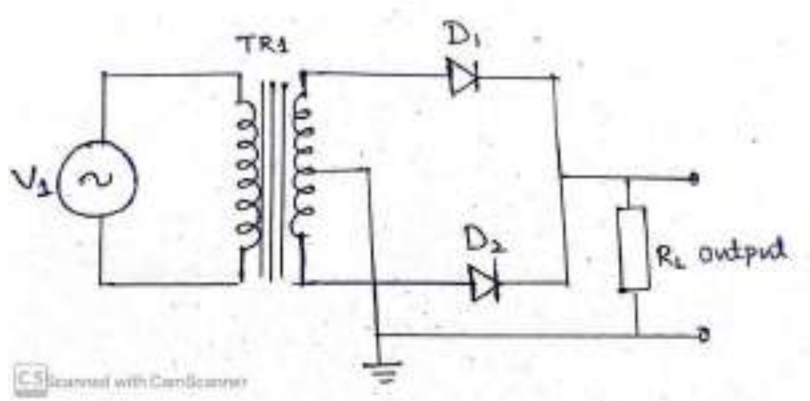
1. Center Tapped Full Wave Rectifier
2. Full Wave Bridge Rectifier

Center Tapped Full Wave Rectifier

Center tap is the contact made at the middle of the winding of the transformer.

In the center tapped full wave rectifier two diodes were used. These are connected to the center tapped secondary winding of the transformer. Above circuit diagram shows the center tapped full wave rectifier. It has two diodes. The positive terminal of two diodes is connected to the two ends of the transformer. Center tap divides the total secondary voltage into equal parts.

Center Tapped Full Wave Rectifier Working*:*



The primary winding of the center tap transformer is applied with the Ac voltage. Thus the two diodes connected to the secondary of the transformer conducts alternatively. For the positive half cycle of the input diode D1 is connected to the positive terminal and D2 is connected to the negative terminal. Thus diode D1 is in forward bias and the diode D2 is reverse biased. Only diode D1 starts conducting and thus current flows from diode and it appears across the load RL. So positive cycle of the input appears at the load.

During the negative half cycle the diode D2 is applied with the positive cycle. D2 starts conducting as it is in forward bias. The diode D1 is in reverse bias and this does not conduct. Thus current flows from diode D2 and hence negative cycle is also rectified, it appears at the load resistor RL.

By comparing the current flow through load resistance in the positive and negative half cycles, it can be concluded that the direction of the current flow is same. Thus the frequency of rectified output voltage is two times the input frequency. The output that is rectified is not pure, it consists of a dc component and a lot of ac components of very low amplitudes.

Peak Inverse Voltage (PIV) of Centre Tap Full Wave Rectifier:

PIV is defined as the maximum possible voltage across a diode during its reverse bias. During the first half that is positive half of the input, the diode D1 is forward bias and thus conducts providing no resistance at all. Thus, the total voltage Vs appears in the upper-half of the ac supply, provided to the load resistance R. Similarly, in the case of diode D2 for the lower half of the transformer total secondary voltage developed appears at the load. The amount of voltage that drops across the two diodes in reverse bias is given as

$$\text{PIV of D2} = Vm + Vm = 2Vm$$

$$\text{PIV of D1} = 2Vm$$

Vm is the voltage developed across upper and lower halves.

Peak Current

The peak current is the instantaneous value of the voltage applied to the rectifier. It can be written as

$$Vs = Vsm \, Sinwt$$

Let us assume that the diode has a forward resistance of RF ohms and a reverse resistance is equal to infinity, thus current flowing through the load resistance RL is given as

$$Im = Vsm \,/\, (R_F + RL)$$

Transformer Utilization Factor:

This can be calculated by considering primary and secondary windings separately. Its value is 0.693.This can be used to determine transformer secondary rating.

Output Current:

Since the current is same through the load resistance R_L in the two halves of the ac cycle, magnitude of dc current I_{dc}, which is equal to the average value of ac current, can be obtained by integrating the current I_1 between 0 and π or current I_2 between π and 2π.

$$I_{dc} = 1/\pi \iint_0^\pi I1(wt) = 1/\pi \iint_0^\pi I_{max} \sin wtd(wt) = 2I_m/\pi$$

DC output voltage:

Average value or dc value of voltage across the load is given by

$$I_{dc} = 1/\pi \iint_0^\pi I1d(wt) = 1/\pi \iint_0^\pi Imax \sin wt(wt) = 2I_m/\pi$$

Root Mean Square (RMS) value of current:

RMS value of current flowing through the load resistance is given as

$$I^2_{RMS} = 1/\pi \iint_0^\pi 12d(wt) = I^2{}_m/2$$

Or

$$I_{rms} = I_m/\sqrt{2}$$

Root Mean Square (RMS) Value of output voltage:

RMS value of voltage across the load is given by:

$$V_{load_{rms}} = I_{rms} * R_{load}$$

Rectification efficiency:

$$P_{dc} = I^2{}_{dc}R_L = 2I_m/\pi = R_L = (4/\pi^2)_m{}^2 R_L$$

As power input to the transformer = power dissipated at the diode + power dissipated at the in load resistance R_L.

$$I_{rms}{}^2R_F + I_{rms}{}^2R_{load} = \{I_m^2/2\}[R_F + R_{load}]$$

Rectification efficiency $\eta = P_{dc}/P_{ac} = ((4/\pi^2)_m{}^2 R_L)/(\{I_m^2/2\}[R_F + R_{load}])$

Ripple factor:

Form factor K_f of the rectified output voltage of a full wave rectifier is given as

$$K_f = I_{rms}/I_{avg} = \frac{I_m/\sqrt{2}}{(2I_m/\pi)} = 1.11$$

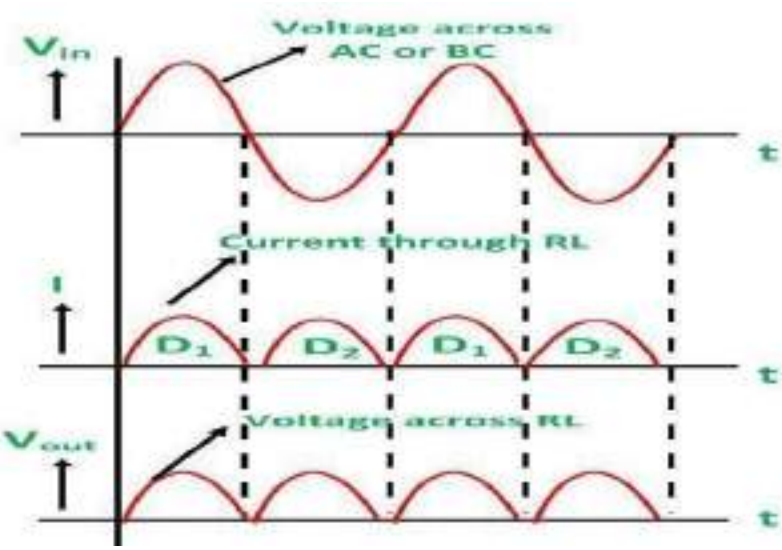
Regulation:

The dc output voltage is given by

$$V_{dc} = I_{dc}R_L = 2/(\pi I_m R_L)$$

$$= 2V_{sm}R_L/(R_F + R_L)$$

$$= (2V_{sm}/\pi) - (I_{dc}R_F)$$



Advantages:-

- ✓ Output is obtained for both cycles of input ac voltages.
- ✓ Efficiency is higher than that of half wave rectifier.

Disadvantages:-

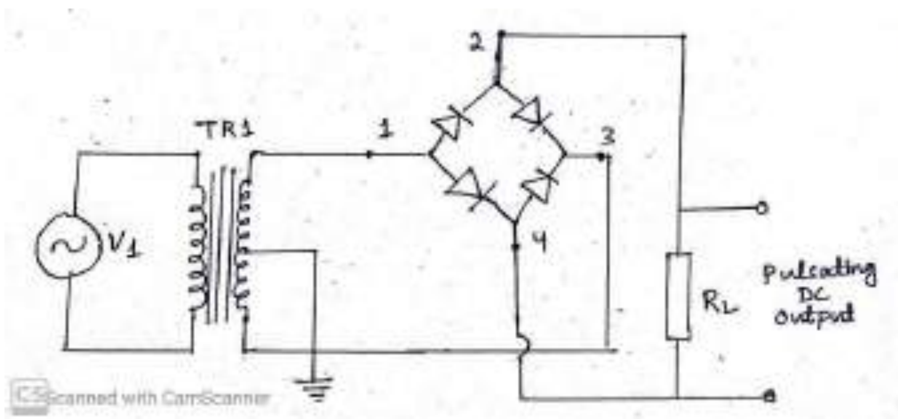
- ✓ Locating center tap on the secondary winding is difficult.
- ✓ The diodes used have high PIV.
- ✓ The d.c output is small as each diode utilizes only one half of the transformer secondary voltage.

FULL WAVE BRIDGE RECTIFIER

Bridge is a type of electrical circuit. Bridge rectifier is a type of rectifier in which diodes were arranged in the form of a bridge. This provides full wave rectification and is of low cost. So it is used in many applications.

Bridge Rectifier:

In bridge rectifier four diodes are used. These are connected as shown in the circuit diagram. The four diodes are connected in the form of a bridge to the transformer and the load as shown.



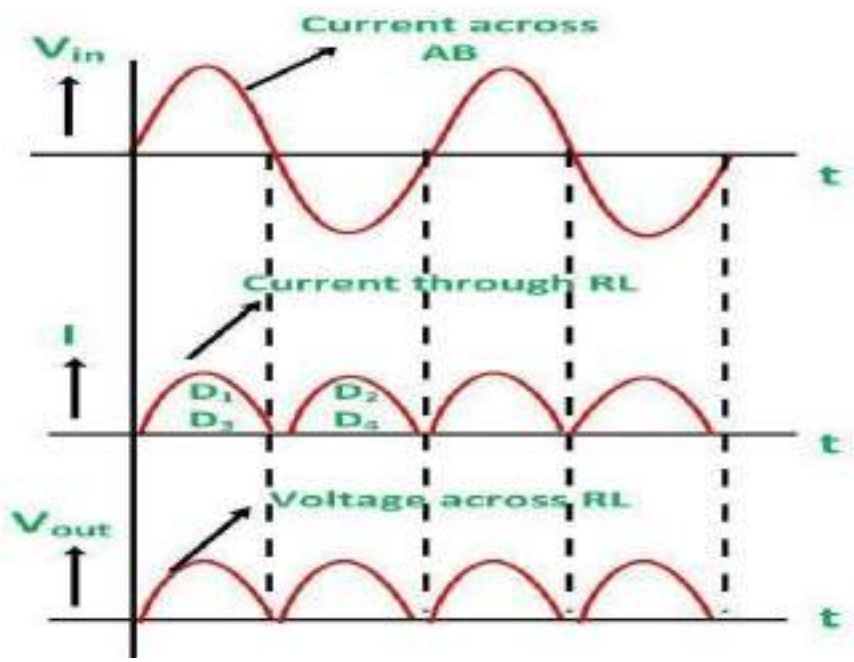
Working of Bridge Rectifier:

The working of a bridge rectifier is simple. The circuit diagram of bridge rectifier is given above. The secondary winding of the transformer is connected to the two diametrically opposite points of the bridge at points 1 and 3. Assume that a load is connected at the output. The load R_{Load} is connected to bridge through points 2 and 4.

During first half cycle of the AC input, the upper portion of the transformer secondary winding is positive with respect to the lower portion. Thus during the first half cycle diodes D1 and D_4 are forward biased. Current flows through path 1-2, enter into the load R_L. It returns back flowing through path 4-3. During this half input cycle, the diodes D_2 and D_3 are reverse biased. Hence there is no current flow through the path 2-3 and 1-4.

During the next cycle lower portion of the transformer is positive with respect to the upper portion. Hence during this cycle diodes D2 and D3 are forward biased. Current flows through the path 3-2 and flows back through the path 4-1.The diodes D1 and D4 are reverse biased. So there is no current flow through the path 1-2 and 3-4.Thus negative cycle is rectified and it appears across the load.

Peak Inverse Voltage (PIV) of a bridge rectifier = Maximum of Secondary Voltage



Advantages:-

- ✓ PIV is one half that of centre tap circuit.
- ✓ Output is twice that of centre tap circuit.
- ✓ Need for centre tapped transformer is eliminated.

Disadvantages:-

- ✓ Requires 4 diodes which increase the cost.

EFFICIENCY-

Let v=Vm sinϴ be the ac voltage to be rectified.

r_f= diode resistance

R_L=load resistance

$$i = \frac{V_m \sin\theta}{r_F + R_L}$$

D.C current:-

$$I_{dc} = Iavg = 2 * \frac{1}{2\pi}\int_0^{\pi} i * d\theta$$

$$= \frac{1}{\pi}\int_0^{\pi} \frac{V_m \sin\theta}{r_F + R_L} * d\Theta$$

$$= \frac{Vm}{\pi(r_F + R_L)}\int_0^{\pi} \sin\theta * d\Theta$$

$$I_{dc} = \frac{V_m}{\pi(r_F + R_L)}[-\cos\Theta]_0$$

$$= \frac{Vm}{\pi(r_F + R_L)}[-\cos\pi + \cos 0]$$

$$= \frac{Vm}{(r_F + R_L)}[-(-1) + 1]$$

$$= \frac{Vm}{(r_F + R_L)} * 2$$

$$= \frac{2I_m}{\pi}$$

AC current:-

$$I^2ac = \frac{1}{\pi}\int_0^{\pi}(Im \sin\theta)^2 * d\theta$$

$$= \frac{1}{}\int_0^{\pi} Im^2 * \sin^2\theta * d\theta$$

$$= \frac{Im^2}{2\pi}\int_0^{\pi}(1 - \cos 2\theta) * d\theta$$

$$I^2_{rms} = \frac{Im^2}{2\pi} * \pi$$

$$I_{rms} = \frac{I_m}{\sqrt{2}}$$

$$P_{ac} = I^2_{rms}(r_f + R_L)$$

$$= \left(\frac{Im^2}{2}\right)^2 (r_f + R_L)$$

Rectifier efficiency (η) = $\frac{P_{dc}}{P_{ac}}$

$$= \frac{(\frac{2Im}{\pi})^2 * R_L}{(\frac{Im}{2})^2 * (r_F + R_L)}$$

$$= \frac{8}{\pi^2 * (\frac{r_F}{R_L} + 1)}$$

As $\frac{r_F}{R_L}$ is a very small value, so it can be neglected from the denominator.

Thus, we get, $\eta = \frac{8}{\pi^2}$

$$\eta \text{ (in \% age)} = \frac{8}{\pi^2} * 100 = 81.2\%$$

RIPPLE FATOR:-

Ripple factor = $\sqrt{(\frac{I_{rms}}{I_{dc}})^2 - 1}$

$$= \sqrt{(\frac{Im/\sqrt{2}}{2Im/\pi})^2 - 1} = 0.48$$

FORM FACTOR:-

✓ It is the ratio of the rms value to the average value.

$$\text{Form factor} = \frac{\text{RMS value}}{\text{Average value}} = \frac{Im/\sqrt{2}}{2Im/\pi} = 1.11$$

PEAK INVERSE VOLTAGE:-

✓ It is the maximum reverse voltage that a diode can withstand without destroying the junction.

$$PIV = 2Vm$$

TRANSFORMER UTILIZATION FACTOR (TUF):-

✓ TUF in a circuit should be as high as possible.

$P_{ac} = V_{rms} I_{rms}$

$= \frac{V_m I_m}{2}$

$P_{dc} = I^2_{dc} . R_L$

$= (\frac{2Im}{\pi})^2 . R_L$

✓ TUF = $\frac{P_{dc}}{P_{ac}}$

$= \frac{(\frac{2I_m}{\pi})^2 R_L}{\frac{V_m I_m}{2}}$

$= \frac{8}{\pi^2}$

$= 0.811$

Bridge Rectifier Circuit Analysis:

In the bridge rectifier circuit, among four diodes two diodes conduct during one half cycles. Thus forward resistance becomes double that is $2R_F$.

Peak Current:

Instantaneous value of the applied voltage to the rectifier is given as

Vs = Vsmax sin wt

Let us assume that the diode has a forward resistance of RF ohms and a reverse resistance is equal to infinity, thus current flowing through the load RL is given as

- i1 = Imax Sin wt and i2 = 0 during the first half cycle and
- i1 = 0 and i2 = Imax sin wt during second half cycle

The total current flowing through the load resistance R_L, the sum of currents i1 and i2 is given as

i = i1 + i2 = Imax sin wt for the complete cycle.

The peak value of current flowing through the load R_L is given as

$$I_{max} = V_{smax} / (2R_F + R_L)$$

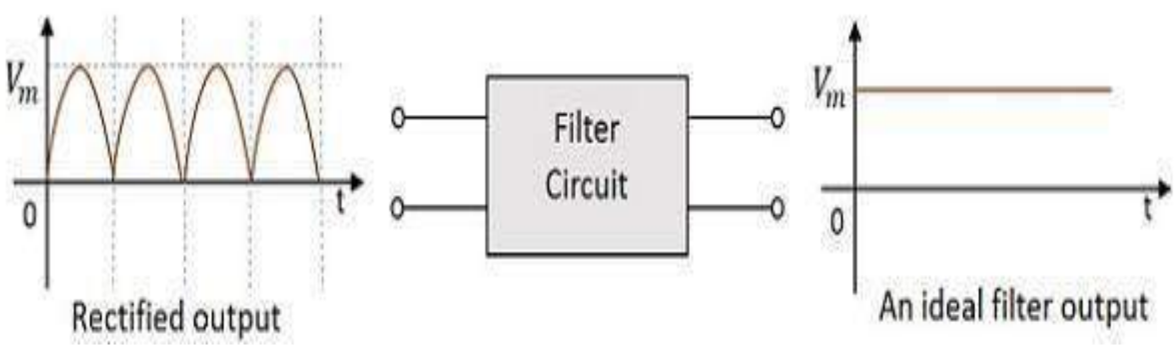
Bridge Rectifier Applications:

- Because of their low cost compared to center tapped they are widely used in power supply circuit.

- This can be used to detect the amplitude of modulated radio signal.
- Bridge rectifiers can be used to supply polarized voltage in welding.

FILTER:-

A filter is a device which removes the ac component of rectifier output but allows the dc component to reach the load. A filter should be connected between the rectifier output & the load. A filter circuit is a combination of inductor & capacitor.



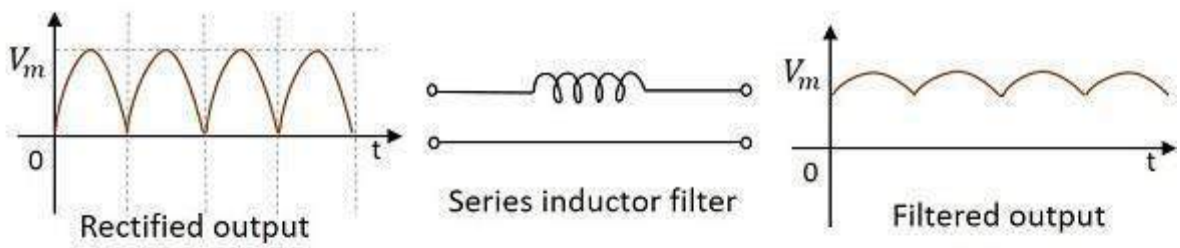
✓ Filter circuits are classified into four types

1. Inductor filter
2. Capacitor filter
3. Choke input filter
4. Capacitor input filter or π filter

INDUCTOR FILTER:-

An inductor filter consists of an inductor which is connected in series, so also named as series inductor filter. The pulsating output of the rectifier is applied across the terminal of the filter circuit. The property of an inductor to block AC and provides zero resistance to DC is used in filtering circuit. When the value of DC output from the rectifier is more than the average value then the rectifier store the excess current in the form of magnetic energy.

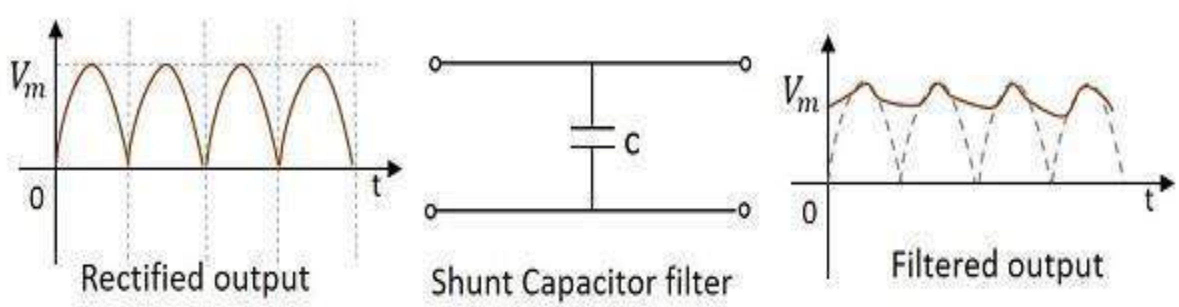
When the value of DC from the rectifier is less than the average value then the inductor release the stored magnetic energy in order to balance the effect of the low value of DC. In this way series inductor filter maintains the regulated DC supply. Moreover, inductor blocks the AC ripples present in the output voltage of rectifier; thus, smooth DC signal can be obtained.



Due to the presence of that leakage A.C component the output is pulsating one.

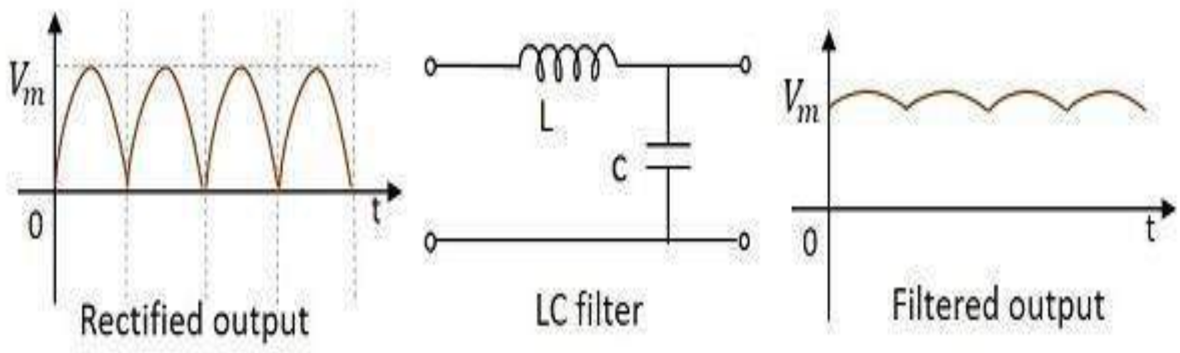
CAPACITOR FILTER:-

A Capacitor filter consists of a capacitor which is connected in parallel with the load. During the rise of voltage cycle it gets charge and this charge is supplied to the load during the fall in the voltage cycle. This process is repeated for each cycle and thus the repel is reduced across the load. It is popular because of its low cost, small size, less weight and good characteristics.



CHOKE INPUT FILTER:-

This filter circuit consists of an inductor L which is connected in series with the rectifier output and a capacitor C which is connected across the load. The inductor has low DC resistance and extremely high AC reactance. Thus, ripples get filtered through choke coil. Some of the residual ripples if present in filtered signal from inductor coil will get bypassed through the capacitor. The reason behind this is that capacitor allows AC and block DC.

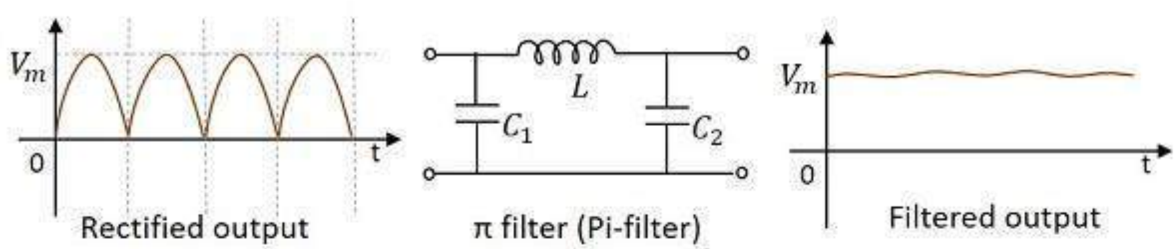


3.3.4. CAPACITOR INPUT FILTER:-

It consists of a capacitor C_1 which is connected across the rectifier output, a choke L in series and a capacitor C_2 which is connected across the load R_L.

The output voltage coming from rectifier also consists of AC components. Thus it is a crucial need to remove these AC ripples to improve the performance of the device. The output from the rectifier is directly applied to the input capacitor. The capacitor provides a low impedance to AC ripples present in the output voltage and high resistance to DC voltage. Therefore, most of the AC ripples get bypassed through the capacitor in input stage only.

The residual AC components which are still present in filtered DC signal gets filtered when they pass through the inductor coil and through the capacitor connected parallel across the load. In this way, the efficiency of filtering increases multiple times.



UNIT-4: TRANSISTORS

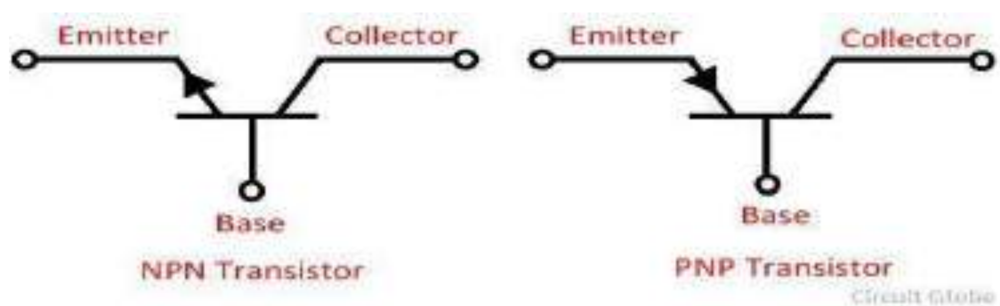
PRINCIPLE OF BIPOLAR JUNCTIONTRANSISTOR:

The transistor is a semiconductor device which transfers a weak signal from low resistance circuit to high resistance circuit. The words trans mean transfer property and istor mean resistance property offered to the junctions. In other words, it is a switching device which regulates and amplifies the electrical signal likes voltage or current.

The transistor consists of two PN diode connected back to back. It has three terminals namely emitter, base and collector. The base is the middle section which is made up of thin layers. The right part of the diode is called emitter diode and the left part is called collector-base diode. These names are given as per the common terminal of the transistor. The emitter based junction of the transistor is connected to forward bias and the collector-base junction is connected in reverse bias which offers a high resistance.

Transistor Symbols

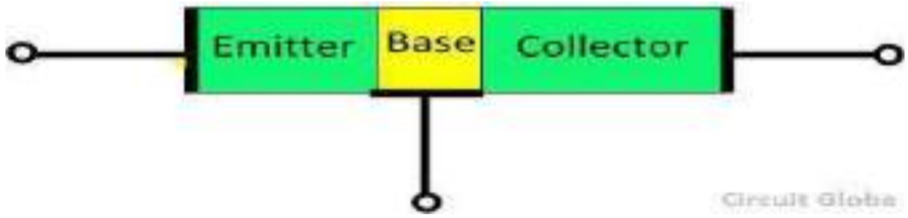
There are two types of transistor, namely NPN transistor and PNP transistor. The transistor which has two blocks of n-type semiconductor material and one block of P-type semiconductor material is known as NPN transistor. Similarly, if the material has one layer of N-type material and two layers of P-type material then it is called PNP transistor. The symbol of NPN and PNP is shown in the figure below.



The arrow in the symbol indicates the direction of flow of conventional current in the emitter with forward biasing applied to the emitter-base junction. The only difference between the NPN and PNP transistor is in the direction of the current.

Transistor Terminals

The transistor has three terminals namely, emitter, collector and base. The terminals of the diode are explained below in details.



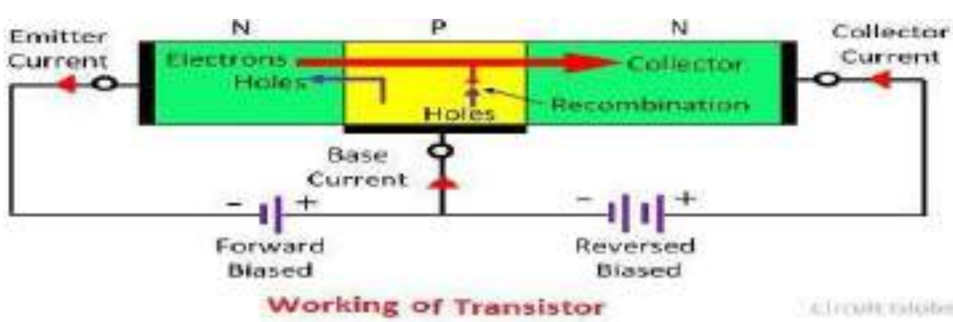
Emitter – The section that supplies the large section of majority charge carrier is called emitter. The emitter is always connected in forward biased with respect to the base so that it supplies the majority charge carrier to the base. The emitter-base junction injects a large amount of majority charge carrier into the base because it is heavily doped and moderate in size.

Collector – The section which collects the major portion of the majority charge carrier supplied by the emitter is called a collector. The collector-base junction is always in reverse bias. Its main function is to remove the majority charges from its junction with the base. The collector section of the transistor is moderately doped, but larger in size so that it can collect most of the charge carrier supplied by the emitter.

Base – The middle section of the transistor is known as the base. The base forms two circuits, the input circuit with the emitter and the output circuit with the collector. The emitter-base circuit is in forward biased and offered the low resistance to the circuit. The collector-base junction is in reverse bias and offers the higher resistance to the circuit. The base of the transistor is lightly doped and very thin due to which it offers the majority charge carrier to the base.

WorkingofTransistor

Usually, silicon is used for making the transistor because of their high voltage rating, greater current and less temperature sensitivity. The emitter-base section kept in forward biased constitutes the base current which flows through the base region. The magnitude of the base current is very small. The base current causes the electrons to move into the collector region or create a hole in the base region.



The base of the transistor is very thin and lightly doped because of which it has less number of electrons as compared to the emitter. The few electrons of the emitter are combined with the hole of the base region and the remaining electrons are moved towards the collector region and constitute the collector current. Thus we can say that the large collector current is obtained by varying the base region.

DIFFERENT MODES OF OPERATION OFTRANSISTOR:

When the emitter junction is in forward biased and the collector junction is in reverse bias, then it is said to be in the active region. The transistor has two junctions which can be biased in different ways. The different working conduction of the transistor is shown in the table below.

CONDITION	EMITTER JUNCTION (EB)	COLLECTOR JUNCTION (CB)	REGION OF OPERATION
FR	Forward-biased	Reversed-biased	Active
FF	Forward-biased	Forward-biased	Saturation
RR	Reversed-biased	Reversed-biased	Cut-off
RF	Reversed-biased	Forward-biased	Inverted

FR – In this case, the emitter-base junction is connected in forward biased and the collector-base junction is connected in reverse biased. The transistor is in the active region and the collector current depends on the emitter current. The transistor, which operates in this region, is used for amplification.

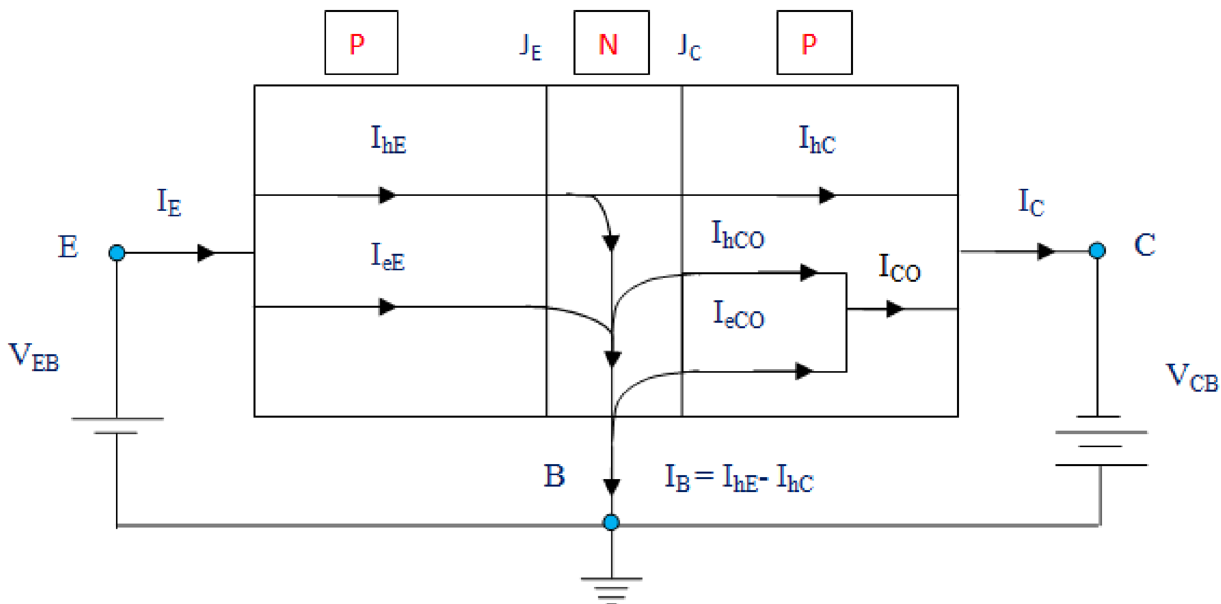
FF – In this condition, both the junction is in forward biased. The transistor is in saturation and the collector current becomes independent of the base current. The transistors act like a closed switch.

RR – Both the current are in reverse biased. The emitter does not supply the majority charge carrier to the base and carriers current are not collected by the collector. Thus the transistors act like a closed switch.

RF – The emitter-base junction is in reverse bias and the collector-base junction is kept in forward biased. As the collector is lightly doped as compared to the emitter junction it does not supply the majority charge carrier to the base. Hence poor transistor action is achieved.

CURRENT COMPONENTS OF IN A TRANSISTOR:

The conduction of current in NPN transistor is owing to electrons and in PNP transistor, it is owing to holes. The direction of current flow will be in opposite direction. Here, we can discuss the current components in a PNP transistor with common base configuration. The emitter-base junction (J_E) is forward biased and the collector-base junction (J_C) is reversed biased as shown in figure. All the current components related to this transistor are shown here.



The current arrives the transistor through the emitter and this current is called emitter current (I_E). This current consists of two constituents – **Hole current** (I_{hE}) and **Electron current** (I_{eE}). I_{eE} is due to passage of electrons from base to emitter and I_{hE} is due to passage of holes from emitter to base.

$$I_E = I_{hE} + I_{eE}$$

Normally, the emitter is heavily doped compared to base in industrial transistor. So, the Electron current is negligible compared to Hole current. Thus we can conclude that, the whole emitter current in this transistor is due to the passage of holes from the emitter to the base.

Some of the holes which are crossing the junction J_E (emitter junction) combines with the electrons present in the base (N-type). Thus, every holes crossing J_E will not arrive at J_C. The remaining holes will reach the collector junction which produces the hole current component, I_{hC}. There will be bulk recombination in the base and the current leaving the base will be

$$I_B = I_{hE} - I_{hC}$$

The electrons in the base which are lost by the recombination with holes (injected into the base across J_E) are refilled by the electrons that enter into the base region. The holes which are arriving at the collector junction (J_C) will cross the junction and it will go into the collector region.

When the emitter circuit is open circuited, then $I_E = 0$ and $I_{hC} = 0$. In this condition, the base and collector will perform as reverse biased diode. Here, the collector current, I_C will be same as reverse saturation current (I_{CO} or I_{CBO}).

I_{CO} is in fact a small reverse current which passes through the PN junction diode. This is due to thermally generated minority carriers which are pushed by barrier potential. This reverse current increase; if the junction is reverse biased and it will have the same direction as the collector current. This current attains a saturation value (I_0) at moderate reverse biased voltage.

When the emitter junction is at forward biased (in active operation region), then the collector current will become

$$I_C = \alpha I_E + I_{CO}$$

The α is the large signal current gain which is a fraction of the emitter current which comprises of I_{hC}.

When the emitter is at closed condition, then $I_E \neq 0$ and collector current will be

$$I_C = I_{CO} + I_{hC}$$

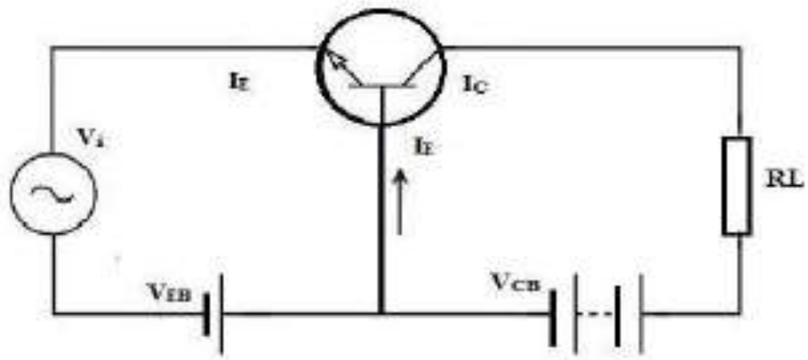
In a PNP transistor, the reverse saturation current (I_{CBO}) will comprises of the current due to the holes passing through the collector junction from the base to collector region (I_{hCO}) and the current due to the electrons which are passing through the collector junction in the opposite direction (I_{eCO}).

$$Therefore,\ I_{CO} = I_{hCO} + I_{eCO}$$

The total current entering into the transistor will be equal to the total current leaving the transistor (according to Kirchhoff's current law).

$$So,\ I_E = I_C + I_B\ or\ I_E = -(I_C + I_B)$$

TRANSISTOR AS AN AMPLIFIER:-



A transistor can be used as **an amplifier** by enhancing the weak signal's strength. With the help of the following transistor amplifier circuit, one can get an idea about how the transistor circuit works as an amplifier circuit.

In the below circuit, the input signal can be applied among the emitter-base junction and the output across the Rc load connected in the collector circuit.

For accurate amplification, always remember that the input is connected in forward-biased whereas the output is connected in reverse-biased. For this reason, in addition to the signal, we apply DC voltage (VEE) in the input circuit as shown in the above circuit.

Generally, the input circuit includes low resistance as a result; a little change will occur in signal voltage at the input which leads to a significant change within the emitter current. Because of the transistor act, emitter current change will cause the same change within the collector circuit.

At present, the flow of collector current through an Rc generates a huge voltage across it. Therefore, the applied weak signal at the input circuit will come out in the amplified form at the collector circuit in the output. In this method, the transistor performs as an amplifier.

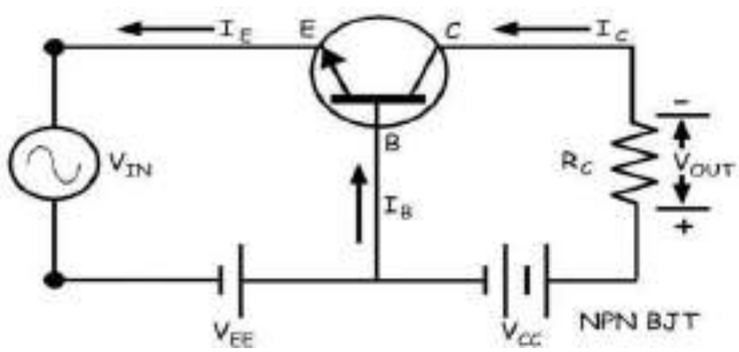
TRANSISTOR CIRCUIT CONFIGURATION & ITSCHARACTERISTICS:

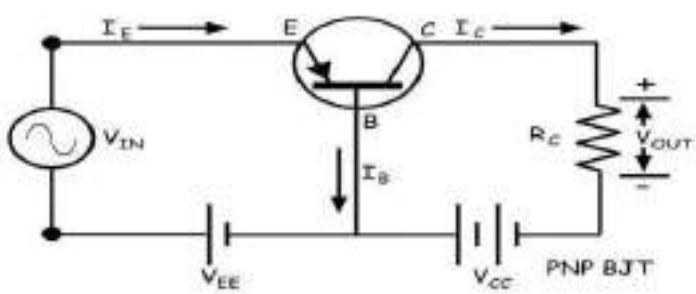
A transistor can be connected in a circuit in the following three ways:

a) common base connection
b) common emitter connection
c) common collector connection

COMMON BASE CONNECTION

Here the base terminal is common to both input and output circuit. The common base configurations or modes are as shown in the figure below. Here, the common base mode of NPN transistor and PNP transistor are shown separately.





Here emitter-base circuit is taken as input circuit and collector base circuit as output circuit.

Current Gain

Here the input current is emitter current I_E and output current is collector current I_C. The current gain is considered as when we only consider the dc biasing voltages

of the circuit and no alternating signal is applied in the input.

$$\alpha_{(dc)} = \frac{I_C}{I_E}$$

Now if we consider the alternating signal applied to the input then the current amplification factor (α) at a constant collector-base voltage, would be

$$\alpha = \frac{\Delta I_C}{\Delta I_E}$$

Here it is seen that neither of current gain and current amplification factor has value more than unity since collector current in no way can be more than emitter current. But as we know that the emitter current and collector current are nearly equal in a bipolar junction transistor, these ratios would be very near to unity. The value generally ranges from 0.9 to even 0.99.

Expression of Collector Current

- If the emitter circuit is open, there will be no emitter current (I_C = 0). But in this condition, there will be a tiny current flowing through the collector region. This is because of flow of minority charge carriers and this is the reverse leakage current.
- As this current flow through collector and base keeping the emitter terminal open, the current is denoted as I_{CBO}. In small power rated transistor the reverse leakage current I_{CBO} is quite small and generally, we neglect it during calculations but in high power rated transistor this leakage current cannot be neglected.
- This current is highly dependent on the temperature so at high temperatures the reverse leakage current I_{CBO} cannot be neglected during calculations.

$$I_C = \alpha I_E + I_{CBO}$$

$$\Rightarrow I_C = \alpha(I_C + I_B) + I_{CBO}$$

$$\Rightarrow I_C(1-\alpha) = \alpha I_B + I_{CBO}$$

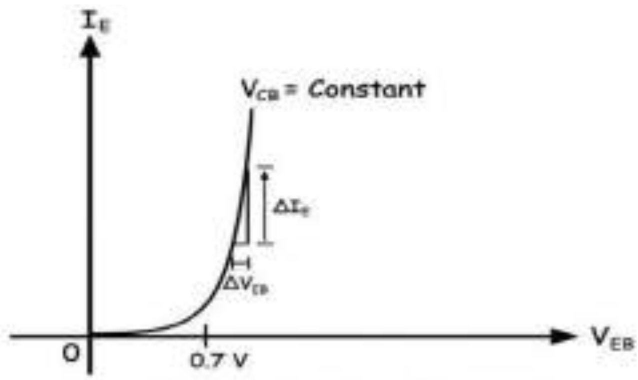
$$\Rightarrow I_C = \frac{\alpha}{1-\alpha} I_B + \frac{1}{1-\alpha} I_{CBO}$$

This expression proves that collector current also depends on base current.

Characteristic of Common Base Connection

Input Characteristic

This is drawn between input current and input voltage of the transistor itself. The input current is emitter current (I_E) and the input voltage is emitter-base voltage (V_{EB}). After crossing emitter-base junction forward barrier potential emitter current (I_E) starts increasing rapidly with increasing emitter-base voltage (V_{EB}).



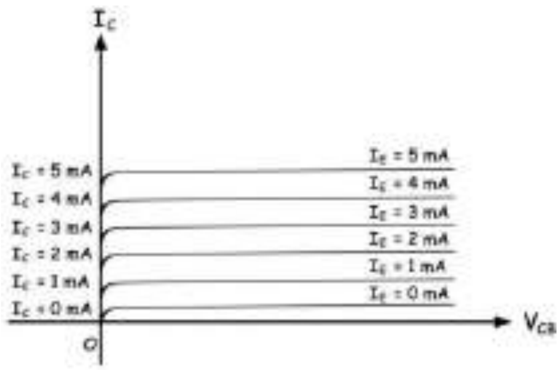
Input Characteristic of Common Base PNP BJT

The input resistance of the circuit is the ratio of change in emitter-base voltage (ΔV_{EB}) to emitter current (ΔI_E) at a constant collector-base voltage (V_{CB} = Constant). As the change in emitter current is quite large compared to the change in emitter-base voltage ($\Delta I_E >> \Delta V_{EB}$), the input resistance of the common base transistor is quite small.

$$r_o = \frac{\Delta V_{CB}}{\Delta I_C} \quad When,\ I_E = Constatnt$$

Output Characteristic

- Collector current gets only constant value when there is sufficient reverse biased established between base and collector region. This is why there is a rise of collector current with an increase of collector-base voltage when this voltage has very low value.
- But after a certain collector-base voltage the collector-base junction gets sufficient reverse biased and hence the collector current becomes constant for a certain emitter current and it entirely depends on the emitter current.

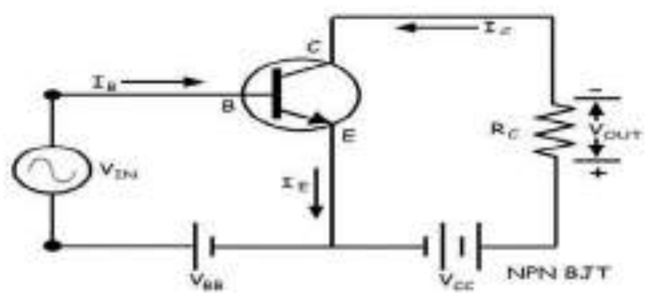


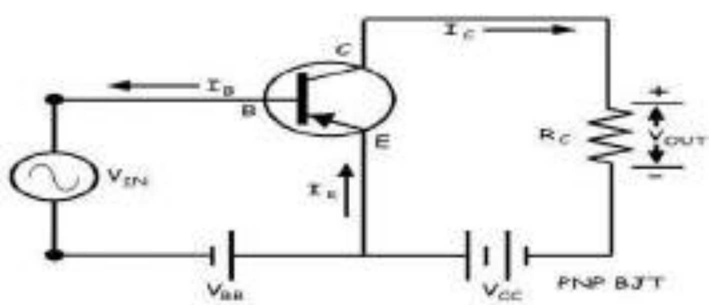
The ratio of change in collector-base voltage to the change in collector current is defined as the output resistance of common base mode of a transistor. Naturally, the value of output resistance is very high in the common base mode of a transistor.

$$r_o = \frac{\Delta V_{CB}}{\Delta I_C} \quad When,\ I_E = Constatnt$$

COMMON EMITTER CONNECTION

Common Emitter Transistor is the most commonly used transistor connection. Here the emitter terminal is common for both input and output circuit. The circuit connected between base and emitter is the input circuit and the circuit connected between collector and emitter is the output circuit. The common emitter mode of NPN transistor and PNP transistor are shown separately in the figure below.





Current Gain

In common emitter configuration, the input current is base current (I_B) and the output current is collector current (I_C). In bipolar junction transistor, the base current controls the collector current. The ratio of change in collector current (ΔI_C) to change in base current (ΔI_B) is defined as the current gain of common emitter transistor. In a bipolar junction transistor, the emitter current (I_E) is the sum of the base current (I_B) and collector current (I_C).

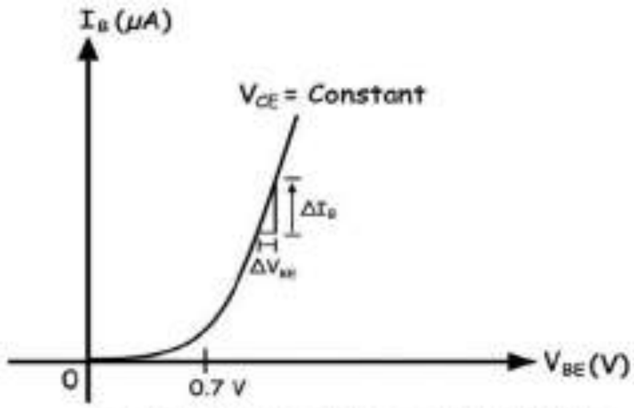
If base current changes, the collector current also changes and as a result the emitter current gets also changed accordingly.

$$\beta = \frac{\Delta I_C}{\Delta I_B}$$

Again the ratio of change of collector current to the corresponding change in emitter current is denoted by α
As the value of base current is quite low compared to the collector current (IB << IC), the current gain in a common emitter transistor is quite high and it ranges from 20 to 500.

Characteristic of Common Emitter Transistor

- In common emitter mode of the transistor, there are two circuits – input circuit and the output circuit. In the input circuit, the parameters are base current and base-emitter voltage.
- The characteristic curve drawn against variations of base current and base-emitter voltage is input characteristic of a common emitter transistor.



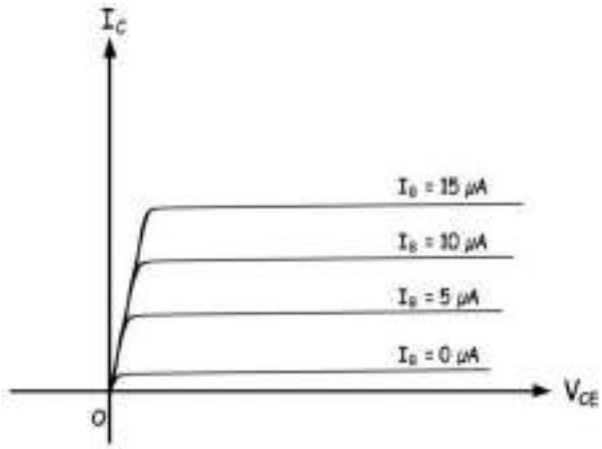
Input Characteristic of Common Emitter NPN BJT

Input resistance of the circuit is:

$$r_i = \frac{\Delta V_{BE}}{\Delta I_B} \quad When,\ V_{CE}\ is\ constant$$

Output Characteristic of Common Emitter Transistor

- The output characteristic is drawn against variations of output current and the output voltage of the transistor. The collector current is output current and collector-emitter voltage is the output voltage of the transistor.

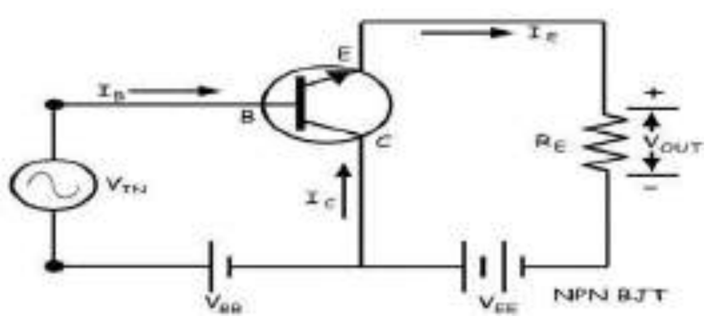


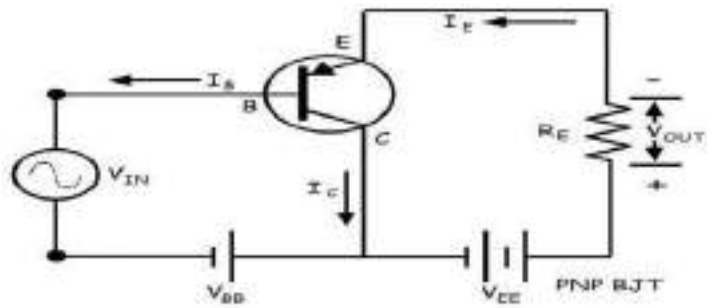
The output resistance would be

$$r_o = \frac{\Delta V_{CE}}{\Delta I_C}\ \ When,\ I_B\ is\ constant$$

COMMON COLLECTOR:

In common collector configuration the input circuit is between base and collector terminal and the output circuit is between emitter and collector terminal.





The ratio of change of emitter current to change of base current is defined as the current gain of common collector configuration. This is denoted as,

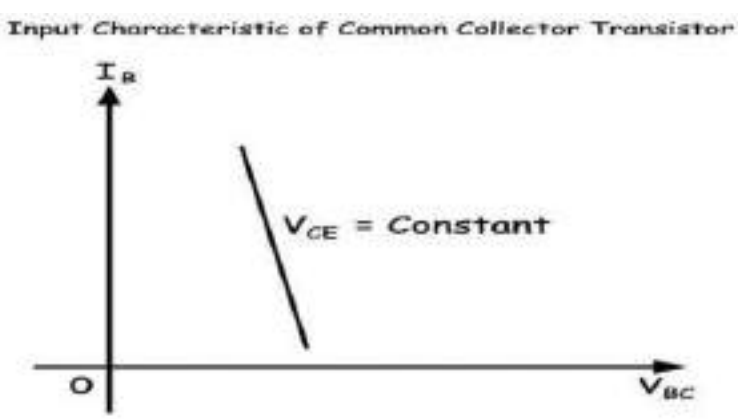
$$\frac{I_E}{I_B}$$

The current amplification factor of the circuit is the ratio of change of emitter current to change of base current when a time-varying signal is applied to the input.

$$\gamma = \frac{\Delta I_E}{\Delta I_B}$$

Input Characteristic of Common Collector Transistor

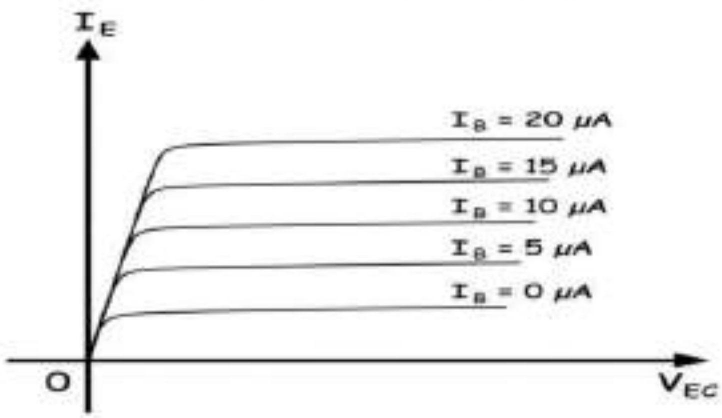
The input current is base current and input voltage of the transistor is base-collector voltage. The base-collector junction is reverse biased and hence with increasing base-collector voltage the reverse biasing of the junction increases. This causes base current to decrease slightly with the increase in base-collector voltage.



Output Characteristic of Common Collector Transistor

The output characteristic of a common collector transistor is nearly the same as the output characteristic of a common emitter transistor. The only difference that here in the case of common collector configuration the output current is emitter current instead of collector current as in the case of common emitter configuration. Here also for a fixed base current, the emitter current increases linearly with increasing collector-emitter voltage up to a certain level of this voltage and then the emitter current gets almost constant irrespective of collector-emitter voltage. Although there would be a very slow increase of emitter current with the collector-emitter voltage as shown in the characteristic curve below.

Output Characteristic of Common Collector Transistor



UNIT-5: TRANSISTOR CIRCUITS

TRANSISTOR BIASING:-

- ✓ Biasing is the process of providing DC voltage which helps in functioning of the circuit.
- ✓ A transistor is biased in order to make the emitter base junction forward biased and collector base junction reverse biased, so that it maintains in active region, to work as an amplifier.
- ✓ The proper flow of zero signal collector current and the maintenance of proper collector emitter voltage during the passage of a signal is known as transistor biasing.
- ✓ The circuit which provides transistor biasing is known as biasing circuit.
- ✓ Transistor biasing is basically classified into 4 types:

  (a) Fixed biasing

  (b) Emitter stabilized biasing

  (c) Voltage divider biasing

  (d) DC biasing with voltage feedback

STABILISATION:

For a transistor to be operated as a faithful amplifier, the operating point should be stabilized. Let us have a look at the factors that affect the stabilization of operating point.

Factors affecting the operating point

The main factor that affects the operating point is the temperature. The operating point shifts due to change in temperature.

As temperature increases, the values of I_{CE}, β, V_{BE} gets affected.

- I_{CBO} gets doubled (for every 10^o rise)
- V_{BE} decreases by 2.5mv (for every 1^o rise)

So the main problem which affects the operating point is temperature. Hence operating point should be made independent of the temperature so as to achieve stability. To achieve this, biasing circuits are introduced.

Stabilization

The process of making the operating point independent of temperature changes or variations in transistor parameters is known as Stabilization.

Once the stabilization is achieved, the values of I_C and V_{CE} become independent of temperature variations or replacement of transistor. A good biasing circuit helps in the stabilization of operating point.

Need for Stabilization

Stabilization of the operating point has to be achieved due to the following reasons.

- Temperature dependence of I_C
- Individual variations
- Thermal runaway

Let us understand these concepts in detail.

- Temperature Dependence of I_C

As the expression for collector current I_C is

$$IC=\beta I_B + I_{CEO}$$

$$=\beta I_B + (\beta+1) I_{CBO}$$

The collector leakage current I_{CBO} is greatly influenced by temperature variations. To come out of this, the biasing conditions are set so that zero signal collector current I_C = 1 mA. Therefore, the operating point needs to be stabilized i.e. it is necessary to keep I_C constant.

- Individual Variations

As the value of β and the value of V_{BE} are not same for every transistor, whenever a transistor is replaced, the operating point tends to change. Hence it is necessary to stabilize the operating point.

- Thermal Runaway

As the expression for collector current I_C is

$$IC=\beta I_B + I_{CEO}$$

$$=\beta I_B + (\beta+1) I_{CBO}$$

The flow of collector current and also the collector leakage current causes heat dissipation. If the operating point is not stabilized, there occurs a cumulative effect which increases this heat dissipation.

The self-destruction of such an un-stabilized transistor is known as Thermal run away.

In order to avoid thermal runaway and the destruction of transistor, it is necessary to stabilize the operating point, i.e., to keep I_C constant.

STABILITY FACTOR:

It is understood that I_C should be kept constant in spite of variations of I_{CBO} or I_{CO}. The extent to which a biasing circuit is successful in maintaining this is measured by Stability factor. It denoted by S.

The rate of change of collector current I_C with respect to the collector leakage current I_{CO} at constant β and I_B is called Stability factor.

$$S = d\,I_c \,/\, d\,I_{co} \text{ at constant } I_B \text{ and } \beta$$

Hence we can understand that any change in collector leakage current changes the collector current to a great extent. The stability factor should be as low as possible so that the collector current doesn't get affected. S=1 is the ideal value.

The general expression of stability factor for a CE configuration can be obtained as under.

$$IC = \beta\, I_B + (\beta+1)\, I_{CO}$$

Differentiating above expression with respect to I_C, we get

$$1 = \beta\, \{d\, I_B / d\, I_C\} + \{(\beta+1)\}\, d\, I_{CO} / d\, I_C$$

Or

$$1 = \beta\, d\, I_B / d\, I_C + (\beta+1)\, S$$

$$\text{Since } d\, I_{CO} / d\, I_C = 1/S$$

Or

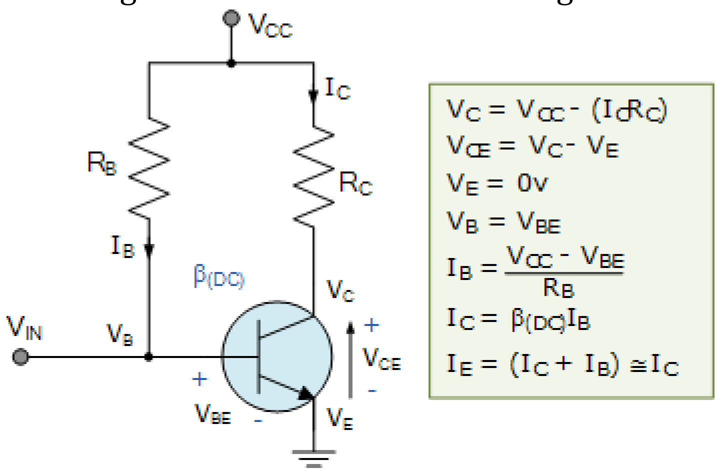
$$S = (\beta+1) / \{1-\beta\, (d\, I_B / d\, I_C)\}$$

Hence the stability factor S depends on β, I_B and I_C.

DIFFERENT METHOD OF TRANSISTORSBIASING:

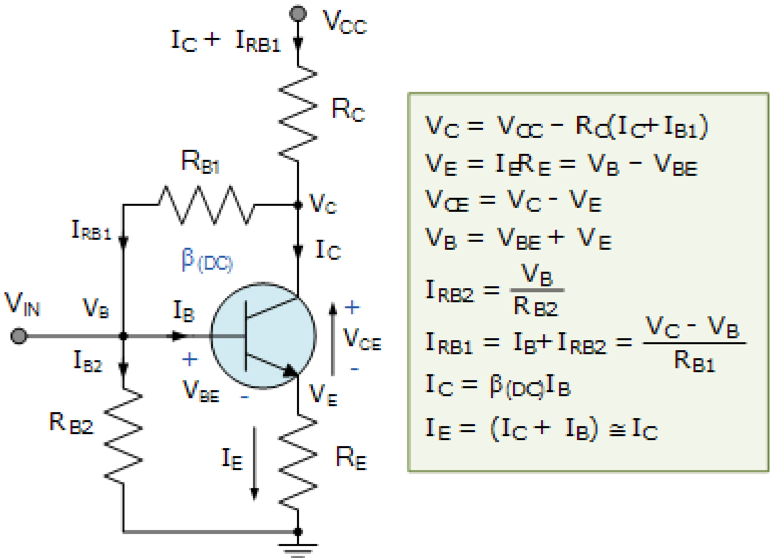
FIXED BIASING:

- ✓ Fixed biasing is also known as base biasing.



- ✓ The above circuit shown is called as a "fixed base bias circuit", because the transistors base current, I_B remains constant for given values of Vcc, and therefore the transistors operating point must also remain fixed.
- ✓ This two resistor biasing network is used to establish the initial operating region of the transistor using a fixed current bias.
- ✓ This type of transistor biasing arrangement is also beta dependent biasing as the steady-state condition of operation is a function of the transistors beta β value, so the biasing point will vary over a wide range for transistors of the same type as the characteristics of the transistors will not be exactly the same.
- ✓ The emitter diode of the transistor is forward biased by applying the required positive base bias voltage via the current limiting resistor R_B.
- ✓ Assuming a standard bipolar transistor, the forward base-emitter voltage drop would be 0.7V. Then the value of R_B is simply: $(V_{CC} - V_{BE})/I_B$ where I_B is defined as I_C/β.
- ✓ With this single resistor type of biasing arrangement the biasing voltages and currents do not remain stable during transistor operation and can vary enormously.
- ✓ Also the operating temperature of the transistor can adversely affect the operating point.

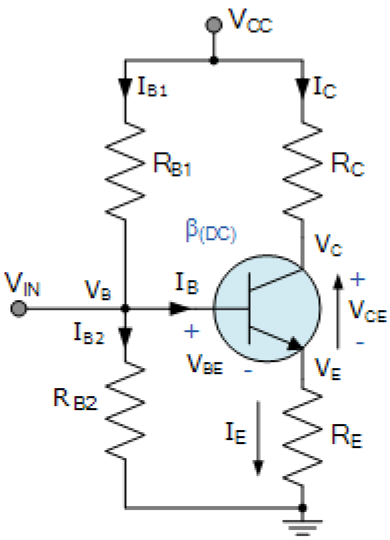
EMITTER STABILIZED BIASING:



- ✓ This type of transistor biasing configuration, often called self-emitter biasing, uses both emitter and base-collector feedback to stabilize the collector current even further.
- ✓ This is because resistors R_{B1} and R_E as well as the base-emitter junction of the transistor are all effectively connected in series with the supply voltage, V_{CC}.
- ✓ The downside of this emitter feedback configuration is that it reduces the output gain due to the base resistor connection.
- ✓ The collector voltage determines the current flowing through the feedback resistor, R_{B1} producing what is called "degenerative feedback".
- ✓ The current flowing from the emitter, I_E (which is a combination of $I_C + I_B$) causes a voltage drop to appear across R_E in such a direction, that it reverse biases the base-emitter junction.
- ✓ So if the emitter current increases, due to an increase in collector current, voltage drop $I*R_E$ also increases. Since the polarity of this voltage reverse biases the base-emitter junction, I_B automatically decrease. Therefore the emitter current increase less than it would have done had there been no self-biasing resistor.
- ✓ Generally, resistor values are set so that the voltage dropped across the emitter resistor R_E is approximately 10% of V_{CC} and the current flowing through resistor R_{B1} is 10% of the collector current I_C.
- ✓ Thus this type of transistor biasing configuration works best at relatively low power supply voltages.

VOLTAGE DIVIDER BIASING:

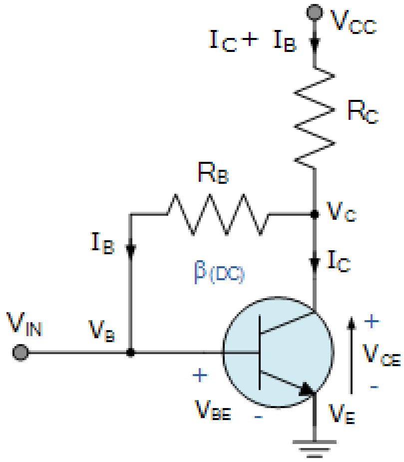
- ✓ Voltage divider biasing is also known as universal biasing.



$$V_C = V_{CC} - R_C I_C = (V_E + V_{CE})$$
$$V_E = I_E R_E = V_B - V_{BE}$$
$$V_{CE} = V_C - V_E = V_{CC} - (I_C R_C + I_E R_E)$$
$$V_B = V_{BE} + V_E = V_{RB2} = \left(\frac{R_{B2}}{R_{B1} + R_{B2}}\right) V_{CC}$$
$$I_{B2} = \frac{V_B}{R_{B2}}$$
$$I_{B1} = I_B + I_{B2} = \frac{V_{CC} - V_B}{R_{B1}}$$
$$R_B = \frac{R_{B1} \times R_{B2}}{R_{B1} + R_{B2}} \qquad I_B = \frac{V_B - V_{BE}}{R_B + (1+\beta) R_E}$$
$$I_C = \beta_{(DC)} I_B$$
$$I_E = I_C + I_B = \frac{V_E}{R_E}$$

- ✓ Here the common emitter transistor configuration is biased using a voltage divider network to increase stability.
- ✓ The name of this biasing configuration comes from the fact that the two resistors R_{B1} and R_{B2} form a voltage or potential divider network across the supply with their center point junction connected the transistors base terminal as shown.
- ✓ This voltage divider biasing configuration is the most widely used transistor biasing method.
- ✓ The emitter diode of the transistor is forward biased by the voltage value developed across resistor R_{B2}.
- ✓ The voltage divider network biasing makes the transistor circuit independent of changes in beta as the biasing voltages set at the transistors base, emitter, and collector terminals are not dependant on external circuit values.
- ✓ To calculate the voltage developed across resistor R_{B2} and the voltage applied to the base terminal we simply use the voltage divider formula for resistors in series.
- ✓ Generally the voltage drop across resistor R_{B2} is much less than for resistor R_{B1}. Clearly the transistors base voltage V_B with respect to ground will be equal to the voltage across R_{B2}.
- ✓ The amount of biasing current flowing through resistor R_{B2} is generally set to 10 times the value of the required base current I_B so that it is sufficiently high enough to have no effect on the voltage divider current or changes in Beta.

DC BIAS WITH COLLECTOR FEEDBACK:



$$V_C = V_{CC} - R_C(I_C + I_B)$$
$$V_E = 0v$$
$$V_B = V_{BE}$$
$$I_B = \frac{V_C - V_B}{R_B}$$
$$I_C = \beta_{(DC)} I_B$$
$$I_E = (I_C + I_B) \cong I_C$$

This self-biasing collector feedback configuration is another beta dependent biasing method which requires two resistors to provide the necessary DC bias for the transistor. The collector to base feedback configuration ensures that the transistor is always biased in the active region regardless of the value of Beta (β). The DC base bias voltage is derived from the collector voltage V_C, thus providing good stability.

In this circuit, the base bias resistor, R_B is connected to the transistors collector C, instead of to the supply voltage rail, Vcc. Now if the collector current increases, the collector voltage drops, reducing the base drive and thereby automatically reducing the collector current to keep the transistors Q-point fixed. Therefore this method of collector feedback biasing produces negative feedback round the transistor as there is a direct feedback from the output terminal to the input terminal via resistor, R_B.

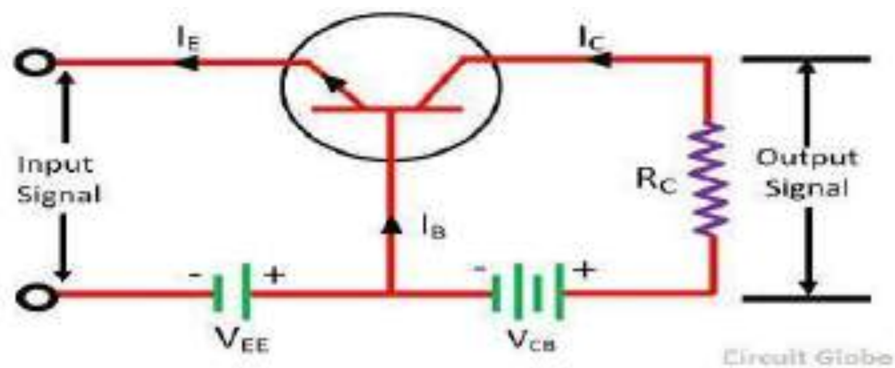
Since the biasing voltage is derived from the voltage drop across the load resistor, R_L, if the load current increases there will be a larger voltage drop across R_L, and a corresponding reduced collector voltage, V_C. This effect will cause a corresponding drop in the base current, I_B which in turn, brings I_C back to normal.

The opposite reaction will also occur when the transistors collector current reduces. Then this method of biasing is called self-biasing with the transistors stability using this type of feedback bias network being generally good for most amplifier designs.

UNIT-6: TRANSISTOR CIRCUITS AND OSCILLATORS:

6.3 PRACTICAL CIRCUIT OF TRANSISTORAMPLIFIER

The transistor raises the strength of a weak signal and hence acts an amplifier. The transistor amplifier circuit is shown in the figure below. The transistor has three terminals namely emitter, base and collector. The emitter and base of the transistor are connected in forward biased and the collector base region is in reverse bias. The forward bias means the P-region of the transistor is connected to the positive terminal of the supply and the negative region is connected to the N-terminal and in reverse bias just opposite of it has occurred.



The input signal or weak signal is applied across the emitter base and the output is obtained to the load resistor R_C which is connected in the collector circuit. The DC voltage V_{EE} is applied to the input circuit along with the input signal to achieve the amplification. The DC voltage V_{EE} keeps the emitter-base junction under the forward biased condition regardless of the polarity of the input signal and is known as a bias voltage.

When a weak signal is applied to the input, a small change in signal voltage causes a change in emitter current (or we can say a change of 0.1V in signal voltage causes a change of 1mA in the emitter current) because the input circuit has very low resistance. This change is almost the same in collector current because of the transmitter action.

In the collector circuit, a load resistor R_C of high value is connected. When collector current flows through such a high resistance, it produces a large voltage drop across it. Thus, a weak signal (0.1V) applied to the input circuit appears in the amplified form (10V) in the collector circuit.

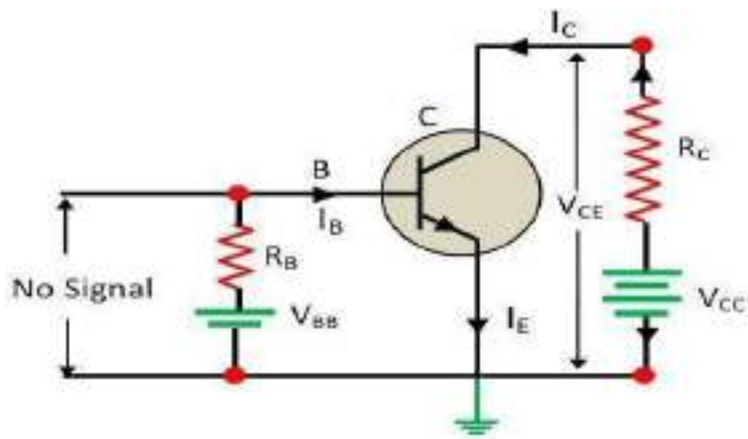
TRANSISTOR LOAD LINE ANALYSIS:

The load line analysis of transistor means for the given value of collector-emitter voltage we find the value of collector current. This can be done by plotting the output characteristic and then determine the collector current I_C with respect to collector-emitter voltage V_{CE}. The load line analysis can easily be obtained by determining the output characteristics of the load line analysis methods.

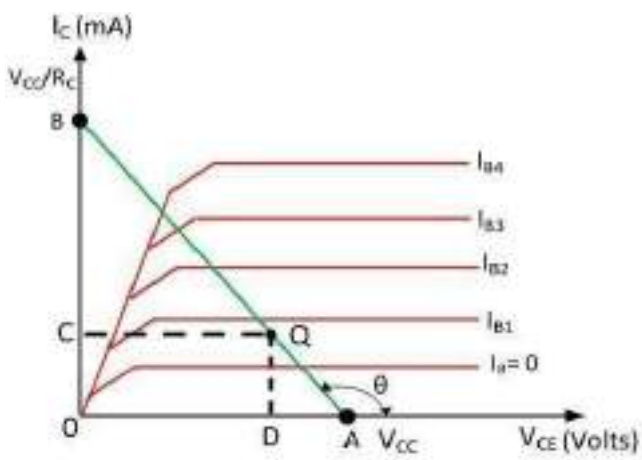
DC LOAD LINE & DC EQUIVALENT CIRCUIT: -

The DC load represents the desirable combinations of the collector current and the collector-emitter voltage. It is drawn when no signal is given to the input, and the transistor becomes bias.

Consider a CE NPN transistor circuit shown in the figure below where no signal is applied to the input side. For this circuit, DC condition will obtain, and the output characteristic of such a circuit is shown in the figure below.



The DC load line curve of the above circuit is shown in the figure below.



By applying Kirchhoff's voltage law to the collector circuit, we get,

$V_{CC} = V_{CE} + I_C R_C$

$V_{CE} = V_{CC} - I_C R_C$ -------- equation 1

The above equation shows that the V_{CC} and R_C are the constant value, and it is the first-degree equation which is represented by the straight line on the output characteristic. This load line is known as a DC load line. The input characteristic is used to determine the locus of V_{CE} and I_C point for the given value of R_C. The end point of the line are located as

1. The collector-emitter voltage V_{CE} is maximum when the collector current I_C = 0 then from the equation (1) we get,

$$V_{CE} = V_{CC} - 0 \, X \, R_C$$
$$V_{CE} = V_{CC}$$

   The first point A (OA = V_{CC}) on the collector-emitter voltage axis shown in the figure above.

2. The collector current I_C becomes maximum when the collector-emitter voltage V_{CE} = 0 then from the equation (1) we get,

   0 = $V_{CC} - \ I_C R_C$

$$I_C = \frac{V_{CC}}{R_C}$$

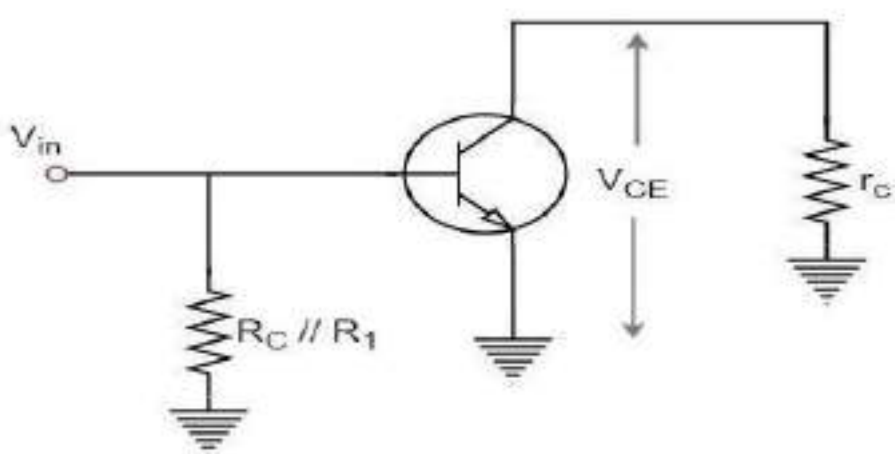
   This gives the second point on the collector current axis as shown in the figure above.

   By adding the points A and B, the DC load line is drawn. With the help of load line, any value of collector current can be determined.

AC LOAD LINE & AC EQUIVALENT CIRCUIT: -

The DC load line discussed previously, analyzes the variation of collector currents and voltages, when no AC voltage is applied. Whereas the AC load line gives the peak-to-peak voltage, or the maximum possible output swing for a given amplifier.

We shall consider an AC equivalent circuit of a CE amplifier for our understanding.



From the above figure,

$$V_{CE} = (\ _C // R_1)\ X\ I_C$$

$$r_C = (\ _C // R_1)$$

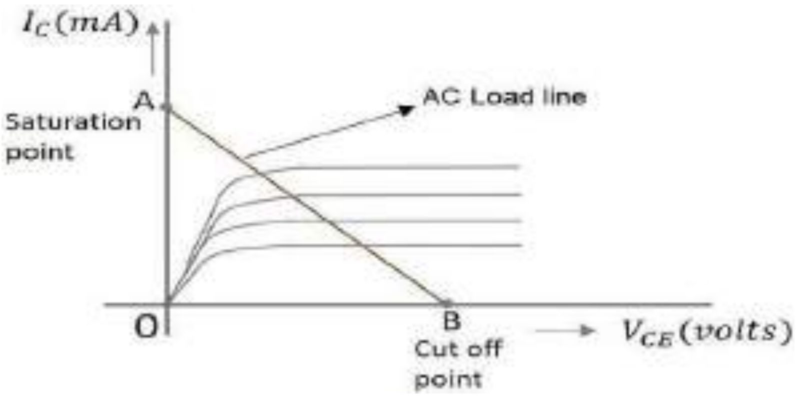
For a transistor to operate as an amplifier, it should stay in active region. The quiescent point is so chosen in such a way that the maximum input signal excursion is symmetrical on both negative and positive half cycles.

Hence,

$$V_{max} = V_{CEQ}\ and\ V_{min} = -V_{CEQ}$$

Where V_{CEQ} is the emitter-collector voltage at quiescent point

The following graph represents the AC load line which is drawn between saturation and cut off points.



From the graph above, the current I_C at the saturation point is

$$I_{C(sat)} = I_{CQ} + ({}^{V_{CEQ}}/_{r_C})$$

The voltage V_{CE} at the cutoff point is

$$V_{CE(off)} = V_{CEQ} + I_{CQ} r_C$$

Hence the maximum current for that corresponding $V_{CEQ} = V_{CEQ}/(R_C//R_1)$ is

$$I_{CQ} = I_{CQ} * (R_C//R_1)$$

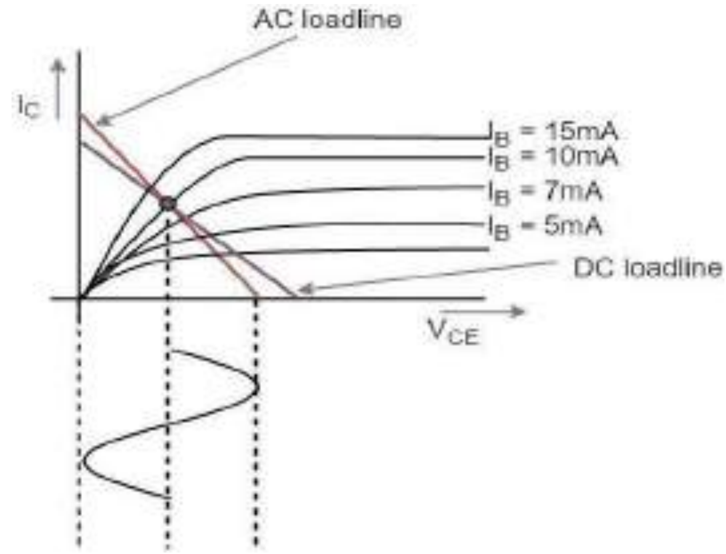
Hence by adding quiescent currents the end points of AC load line are

$$I_{(sat)} = I_{CQ} + V_{CEQ}/(R_C//R_1)$$

$$V_{C(off)} = V_{CEQ} + I_{CQ} * (R_C//R_1)$$

AC and DC Load Line

When AC and DC Load lines are represented in a graph, it can be understood that they are not identical. Both of these lines intersect at the Q-point or quiescent point. The endpoints of AC load line are saturation and cut off points. This is understood from the figure below.



From the above figure, it is understood that the quiescent point (the dark dot) is obtained when the value of base current IB is 10mA. This is the point where both the AC and DC load lines intersect.

PHASE REVERSAL:

The phase relationship between the input and output voltages can be determined by considering the effect of a positive half cycle and negative half cycle separately. Consider the positive half cycle of input signal in which terminal A is positive w.r.t B. Due to this, two voltages, ac and dc will be adding each other, increasing forward bias on base emitter junction. This increases base current. The collector current is β times the base current; hence the collector current will also increase. This increases the voltage drop across R_C. Since $V_C = V_{CC} - I_C R_C$, the increases in I_C results in a drop in collector voltage V_C, as V_{CC} is constant. Thus, as V_i increases in a positive direction, V_o goes in a negative direction and we get negative half cycle of output voltage for positive half cycle at the input.

In the negative half cycle of input, in which terminal A becomes negative w.r.t. terminal B, the ac and dc voltages will oppose each other, reducing forward bias on base-emitter p-n junction. This reduces base current. Accordingly collector current and drop across R_C both reduce, increasing the output voltage. Thus, we get positive half cycle at the output for negative half cycle at the input. Therefore, we can say that there is a phase shift of 180° between input and output voltages for a common emitter amplifier.

HYBRID PARAMETERS OR H PARAMETERS

Hybrid parameters (also known as h parameters) are known as 'hybrid' parameters as they use Z parameters, Y parameters, voltage ratio, and current ratios to represent the relationship between voltage and current in a two port network.

H parameters are useful in describing the input-output characteristics of circuits where it is hard to measure Z or Y parameters (such as a transistor). H parameters encapsulate all the important linear characteristics of the circuit, so they are very useful for simulation purposes. The relationship between voltages and current in h parameters can be represented as:

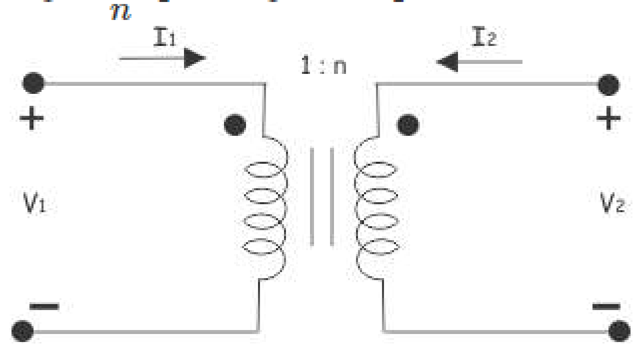
$$V_1 = h_{11} I_1 + h_{12} V_2$$
$$I_2 = h_{21} I_1 + h_{22} V_2$$

This can be represented in matrix form as:

$$\begin{bmatrix} V_1 \\ I_2 \end{bmatrix} = \begin{bmatrix} h_{11} & h_{12} \\ h_{21} & h_{22} \end{bmatrix} \begin{bmatrix} I_1 \\ V_2 \end{bmatrix}$$

To illustrate where h parameters are useful, take the case of an ideal transformer, where Z parameters cannot be used. Since here, the relations between voltages and current in that ideal transformer would be,

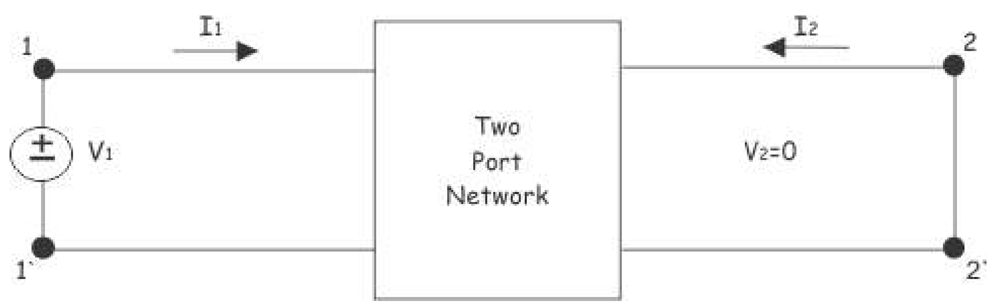
$$V_1 = \frac{1}{n} V_2 \ and \ I_1 = -nI_2$$



Since, in an ideal transformer voltages cannot be expressed in terms of current, it is impossible to analyze a transformer with Z parameters because a transformer does not have Z parameters. The problem can be solved by using hybrid parameters (i.e. h parameters).

Determining h Parameters

Let us short circuit the output port of a two port network as shown below,



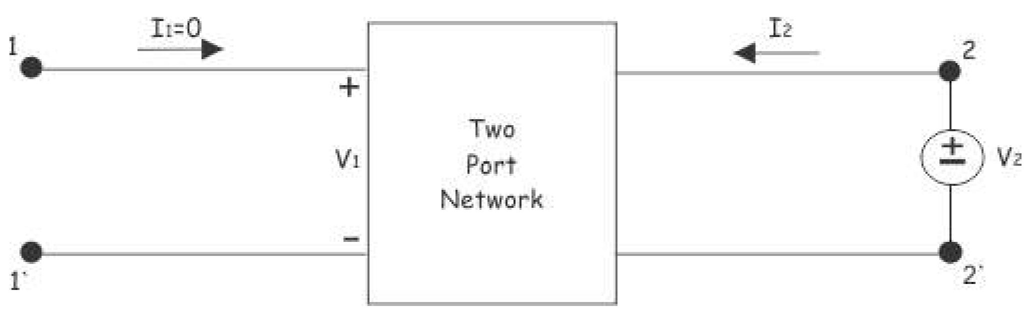
Now, ratio of input voltage to input current, at short circuited output port is:

$$\left.\frac{V_1}{I_1}\right|_{V_2=0} = h_{11}$$

This is referred to as the short circuit input impedance. Now, the ratio of the output current to input current at the short-circuited output port is:

$$\left.\frac{I_2}{I_1}\right|_{V_2=0} = h_{21}$$

This is called short-circuit current gain of the network. Now, let us open circuit the port 1. At that condition, there will be no input current (I_1=0) but open circuit voltage V_1 appears across the port 1, as shown below:



Now:

$$\left.\frac{V_1}{V_2}\right|_{I_1=0} = h_{12} = open\ circuit\ reverse\ voltage\ gain$$

This is referred as reverse voltage gain because, this is the ratio of input voltage to the output voltage of the network, but voltage gain is defined as the ratio of output voltage to the input voltage of a network.

Now:

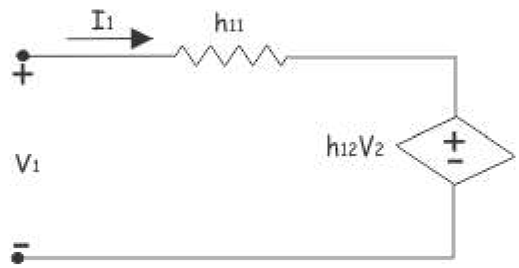
$$\left.\frac{I_2}{V_2}\right|_{I_1=0} = h_{21}$$

It is referred as open circuit output admittance.

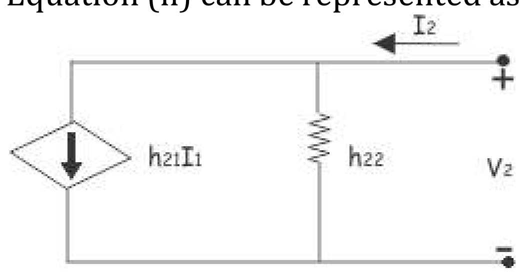
To draw h parameter equivalent network of a two port network, first we have to write the equation of voltages and currents using h parameters. These are:

$$V_1 = h_{11}I_1 + h_{12}V_2 \cdots\cdots\cdots\cdots\cdots (i)$$
$$I_2 = h_{21}I_1 + h_{22}V_2 \cdots\cdots\cdots\cdots\cdots (ii)$$

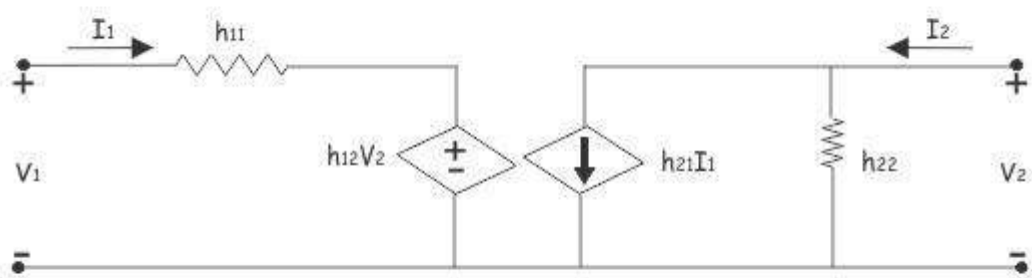
Equation (i) can be represented as a circuit based on Kirchhoff Voltage Law:



Equation (ii) can be represented as a circuit based on Kirchhoff Current Law:



Combining these two parts of the network we get:



The h parameters equivalent network of a two port network

GENERALISED APPROXIMATE MODEL:

In the analysis of transistor amplifier, we have as far used the exact h-model for the transistor. In practice, we may conveniently use an approximately h-model for the transistor which introduces error < 10% in most cases.

This much error may be conveniently tolerated since the h-parameters themselves are not steady but vary considerably for the same type of transistor. We first derive this approximate CE h-model.

Figure 1 gives the equivalent circuit of CE amplifier using exact h-model for CE transistor.

The following steps are used to driving the approximate h-model:

1. If $R_L < 0.1 \frac{1}{h_{oe}}$ and h_{oe}. $R_L < 0.1$, then we may neglected, $\frac{1}{h_{oe}}$ being in parallel with R_L.

2. Having neglected h_{oe}, the collected current I_C equals h_{fe}. I_b and the magnitude of the dependent voltage generator in the emitter circuit is then given by,

$$h_{re} * |V_C| = h_{re} * I_C * R_L \approx h_{re} * h_{fe} * I_b * R_L \text{ ....(1)}$$

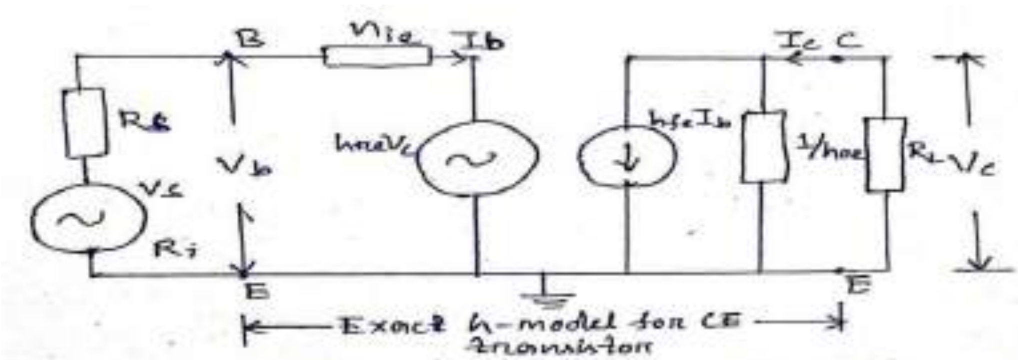


Figure 1: Equivalent circuit using exact h model

But $h_{re} * h_{fe} \approx 0$. Hence the voltage $h_{re} |V_C|$ in the emitter circuit may be neglected in comparison with the voltage drop $h_{ie}.I_b$ provided that R_L is not very large. Then the approximate CE h-model reduces to the form shown in Figure 2.

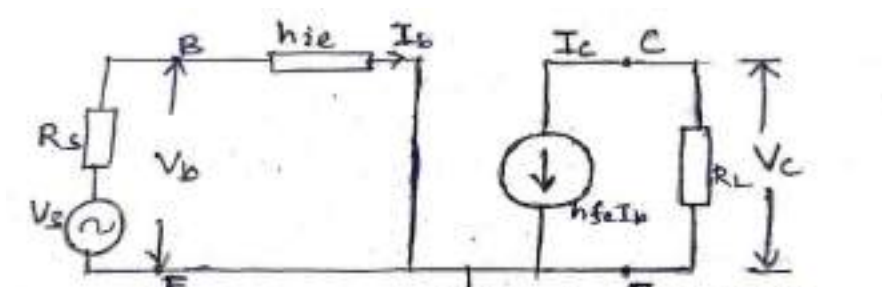


figure-2: Equivalent circuit using approximate h model of CE amplifier

ANALYSIS OF CB, CE, CC AMPLIFIER USING GENERALISED APPROXIMATE MODEL:

Approximate h-model Valid for all the three Configuration

The approximate CE h-model of Figure 2 is redrawn in figure 3. This model may be used for any of the three configurations by grounding the appropriate node and analysis done accordingly. It may be proved that the error in values of A_I, R_i, A_V or output terminal resistance R_{ot} (= $R_0 || R_L$) caused by use of approximate model does not exceed 10% if .$h_{oe} * R_L < 0.1$.

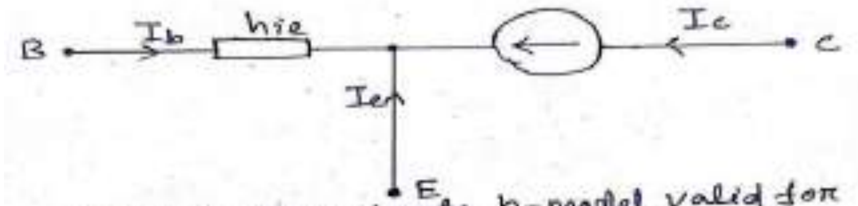


Figure 3: Approximate h-model valid for all the three configurations of a transistor.

Analysis of CE Amplifier using approximate h-model

Figure 2 gives the equivalent circuit of CE amplifier using approximate h-model for the transistor. For this equivalent circuit we get,

Current gain $A_I = \frac{-h_{fe} X I_b}{I_b} = -h_{fe}$ .....(2)

Input resistance $R_i = h_{ie}$

Voltage gain $A_V = A_I \; x \; \frac{RL}{R_i} = \frac{-h_{fe} X R_L}{h_{ie}}$ ......(3)

Output resistance R_0: From this approximate equivalent circuit of figure 1(b) with $V_s = 0$ and with external voltage source connected across the output, we get $I_b = 0$ and therefore $I_C = 0$. Hence output resistance $R_0 = \infty$. However, in actual practice, R_0 lies between $40k\Omega$ *and* $80k\Omega$ depending on the value of R_S.

With load resistance $R_L = 4k\Omega$(the maximum practical value), the output terminal resistance $R_t = R_L \parallel \infty = R_L = 4k\Omega$.

Condition $h_{oe} * R_L < 0.1$. For a typical transistor $h_{oe} = 25 * 10^{-6}$S. Hence to meet the condition that, $h_{|oe|} * R_L < 0.1$.we must use R_L less than $4k\Omega$.

Analysis of CB Amplifier using the Approximate Model

From figure 4 gives the equivalent circuit of a CB amplifier using the approximate model for the transistor as given in figure 2 with base grounded, the input applied between emitter and base and output obtained across load resistor R_L between the collector and the base.

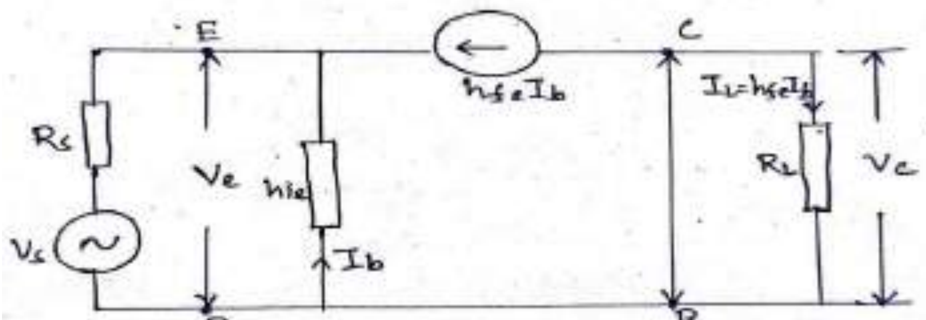


figure-4: Equivalent circuit of a CB amplifier using approximate h-model

Current gain $A_I = \frac{I_L}{I_e} = \frac{-h_{fe} X I_b}{-(1+ h_{fe}) X I_b} = \frac{h_{fe}}{1+h_{fe}}$.....(4)

Input resistance R_I: from figure 4,

$V_e = -I_b * h_{ie}$...... (5)

$I_e = -(1 + h_{fe}) * I_b$ ....(6)

Hence,$_i$ $= \frac{Ve}{I_e} = \frac{-Ib X hie}{-I_b (1+ h_{fe})} = \frac{hie}{1+ h_{fe}}$.....(7)

Voltage Gain A_V: From figure 4,

$$V_C = -h_{fe} * I_b * R_L$$

Hence, $A_V = \frac{C}{V_e} = \frac{-h_{fe} X I_b X R_L}{- I_b X h_e} = \frac{h_{fe} X R_L}{h_{ie}}$.....(8)

Output resistance in the equivalent circuit of figure 3, with V_s = 0, we get I_e = 0. Hence, I_b = 0. Hence the output resistance $R_0 = \infty$.

Output Terminal Resistance $R_{ot} = R_o \| R_L = \infty \| R_L = R_L$ .....(9)

Analysis of CC Amplifier (Emitter Follower) using approximate h-model

Figure 5 gives the equivalent circuit of an emitter follower using the approximate model as given in figure 3, with collector grounded, input signal applied between the base and the ground and the load impedance R_L connected between emitter and ground.

Current gain A_I : from the circuit of figure 5,

Load current $I_L = (1 + h_{fe})I_b$.....(10)

Hence Current gain $A_I = \frac{I_L}{I_b} = (1 + h_{fe})$.....(11)

Input resistance R_i: from figure 5,

$V_b = I_b \, x \, h_{ie} + (1 + h_{fe})_b \, x \, R_L$..... (12)

Hence, $R_i = \frac{Vb}{I_b} = h_{ie} + (1 + h_{fe})_L$...... (13)

Voltage Gain A_V: From figure 5,

$V_e = (1 + h_{fe})_b * R_L$ ...... (14)

Hence,

$$A_V = \frac{e}{V_b} = \frac{(1+h_{fe})I_b X R_L}{I_b X h_{ie}+(1+ h_{fe})I_b X R_L}$$

$$= \frac{(1+h_{fe})R_L}{h_{ie}+(1+h_{fe})R_L} = 1 - \frac{h_{ie}}{h_{ie}+(1+h_{fe})R_L} = 1 - \frac{h_{ie}}{R_i}$$

Output Resistance from figure 5, Open circuit output voltage = V_S

Short circuit output current

$$I = (1 + h_{fe})\, I_b = \frac{(1+h_{fe})V_s}{h_{ie}+R_s}$$

Hence output impedance

$$R_0 = \frac{Open\ circuit\ output\ voltage}{Short\ circuit\ output\ current} = \frac{h_{ie} + R_s}{1 + h_{fe}}$$

Output terminal Impedance $R_{ot} = R_0 \parallel R_L$

Expressions for current gain etc. for the three configurations using approximate h-model.

Expressions for A_I, R_i, A_V, R_0 and R_{ot} using Approximate h-model			
Quantity	CE	CB	CC
A_I	$-h_{fe}$	$\frac{h_{fe}}{1+h_{fe}}$	$1 + h_{fe}$
R_i	h_{ie}	$\frac{h_{ie}}{1+h_{fe}}$	$h_{ie} + (1 + h_{fe})R_L$
A_V	$\frac{h_{fe} \times R_L}{h_{ie}}$	$\frac{h_{fe} \times R_L}{R_e}$	$1 - \frac{h_{ie}}{R_i}$
R_0	∞	∞	$\frac{h_{ie} + R_s}{1 + h_{fe}}$
R_{ot}	R_L	R_L	$R_0 \parallel R_L$

MULTISTAGE TRANSISTOR AMPLIFIERS:

In Multi-stage amplifiers, the output of first stage is coupled to the input of next stage using a coupling device. The following figure shows a two-stage amplifier connected in cascade.

If there are n numbers of stages, the product of voltage gains of those n stages will be the overall gain of that multistage amplifier circuit.

COUPLING:-

Coupling is a process in which the output of one stage is fed as input to the next stage. The main purpose of coupling is to:-

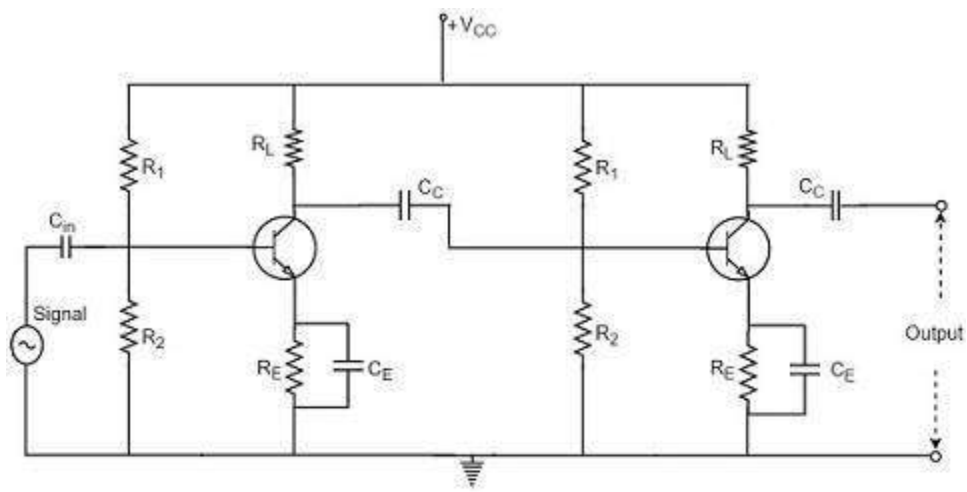
1) Transfer output of one stage to the input of next stage.
2) Isolate the dc condition of one stage from the next stage.

TYPES OF COUPLING:

Coupling is classified into three types. They are:-

- RC coupling
- Transformer coupling
- Direct coupling

RC COUPLED TRANSISTOR AMPLIFIER:-

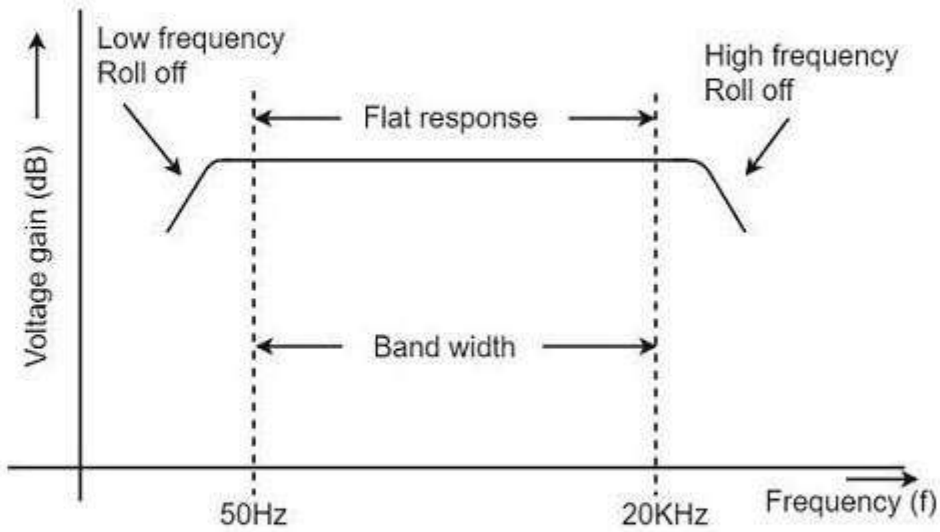


Operation of RC Coupled Amplifier

When an AC input signal is applied to the base of first transistor, it gets amplified and appears at the collector load R_L which is then passed through the coupling capacitor C_C to the next stage. This becomes the input of the next stage, whose amplified output again appears across its collector load. Thus the signal is amplified in stage by stage action.

Frequency Response of RC Coupled Amplifier

Frequency response curve is a graph that indicates the relationship between voltage gain and function of frequency. The frequency response of a RC coupled amplifier is as shown in the following graph.



From the above graph, it is understood that the frequency rolls off or decreases for the frequencies below 50Hz and for the frequencies above 20 KHz. whereas the voltage gain for the range of frequencies between 50Hz and 20 KHz is constant.

We know that,

$$XC=1/2\pi fc$$

It means that the capacitive reactance is inversely proportional to the frequency.

At Low frequencies (i.e. below 50 Hz)

The capacitive reactance is inversely proportional to the frequency. At low frequencies, the reactance is quite high. The reactance of input capacitor C_{in} and the coupling capacitor C_C are so high that only small part of the input signal is allowed. The reactance of the emitter by pass capacitor C_E is also very high during low frequencies. Hence it cannot shunt the emitter resistance effectively. With all these factors, the voltage gain rolls off at low frequencies.

At High frequencies (i.e. above 20 KHz)

Again considering the same point, we know that the capacitive reactance is low at high frequencies. So, a capacitor behaves as a short circuit, at high frequencies. As a result of this, the loading effect of the next stage increases, which reduces the voltage gain. Along with this, as the capacitance of emitter diode decreases, it increases the base current of the transistor due to which the current gain (β) reduces. Hence the voltage gain rolls off at high frequencies.

At Mid-frequencies (i.e. 50 Hz to 20 KHz)

The voltage gain of the capacitors is maintained constant in this range of frequencies, as shown in figure. If the frequency increases, the reactance of the capacitor C_C decreases which tends to increase the gain. But this lower capacitance reactive increases the loading effect of the next stage by which there is a reduction in gain.

Due to these two factors, the gain is maintained constant.

Advantages of RC Coupled Amplifier

The following are the advantages of RC coupled amplifier.

- The frequency response of RC amplifier provides constant gain over a wide frequency range, hence most suitable for audio applications.
- The circuit is simple and has lower cost because it employs resistors and capacitors which are cheap.
- It becomes more compact with the upgrading technology.

Disadvantages of RC Coupled Amplifier

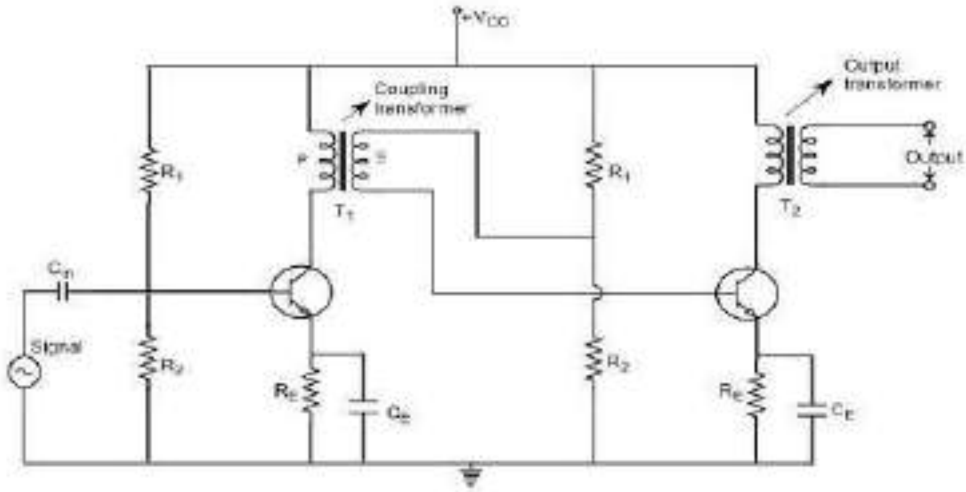
The following are the disadvantages of RC coupled amplifier.

- The voltage and power gain are low because of the effective load resistance.
- They become noisy with age.
- Due to poor impedance matching, power transfer will be low.

TRANSFORMER COUPLED AMPLIFIER:-

In a transformer-coupled amplifier, the stages of amplifier are coupled using a transformer. Let us go into the constructional and operational details of a transformer coupled amplifier.

The figure below shows the circuit diagram of transformer coupled amplifier.



Operation of Transformer Coupled Amplifier

When an AC signal is applied to the input of the base of the first transistor then it gets amplified by the transistor and appears at the collector to which the primary of the transformer is connected.

The transformer which is used as a coupling device in this circuit has the property of impedance changing, which means the low resistance of a stage (or load) can be reflected as a high load resistance to the previous stage. Hence the voltage at the primary is transferred according to the turns ratio of the secondary winding of the transformer.

This transformer coupling provides good impedance matching between the stages of amplifier. The transformer coupled amplifier is generally used for power amplification.

Frequency Response of Transformer Coupled Amplifier

The figure below shows the frequency response of a transformer coupled amplifier. The gain of the amplifier is constant only for a small range of frequencies. The output voltage is equal to the collector current multiplied by the reactance of primary.

At low frequencies, the reactance of primary begins to fall, resulting in decreased gain. At high frequencies, the capacitance between turns of windings acts as a bypass condenser to reduce the output voltage and hence gain.

So, the amplification of audio signals will not be proportionate and some distortion will also get introduced, which is called as Frequency distortion.

Advantages of Transformer Coupled Amplifier

The following are the advantages of a transformer coupled amplifier −

- An excellent impedance matching is provided.
- Gain achieved is higher.
- There will be no power loss in collector and base resistors.
- Efficient in operation.

Disadvantages of Transformer Coupled Amplifier

The following are the disadvantages of a transformer coupled amplifier −

- Though the gain is high, it varies considerably with frequency. Hence a poor frequency response.
- Frequency distortion is higher.
- Transformers tend to produce hum noise.
- Transformers are bulky and costly.

FEEDBACK IN AMPLIFIER:

Feedback is the process by which a fraction of the output signal, either a voltage or a current, is used as an input.

GENERAL THEORY OF FEEDBACK:

The process of injecting a fraction of output energy of some device back to the input is known as Feedback. It has been found that feedback is very useful in reducing noise and making the amplifier operation stable.

Depending upon whether the feedback signal aids or opposes the input signal, there are two types of feedbacks used.

a) Positive Feedback

The feedback in which the feedback energy i.e., either voltage or current is in phase with the input signal and thus aids it is called as Positive feedback.

Both the input signal and feedback signal introduces a phase shift of 180° thus making a 360° resultant phase shift around the loop, to be finally in phase with the input signal.

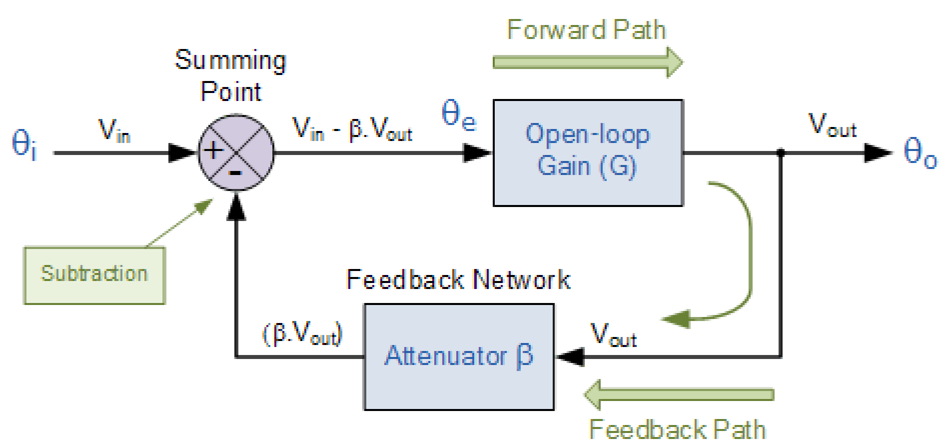
Though the positive feedback increases the gain of the amplifier, it has the disadvantages such as

- Increasing distortion
- Instability

It is because of these disadvantages the positive feedback is not recommended for the amplifiers. If the positive feedback is sufficiently large, it leads to oscillations, by which oscillator circuits are formed. This concept will be discussed in OSCILLATORS tutorial.

b) Negative Feedback

The feedback in which the feedback energy i.e., either voltage or current is out of phase with the input and thus opposes it, is called as negative feedback.



In negative feedback, the amplifier introduces a phase shift of 180° into the circuit while the feedback network is so designed that it produces no phase shift or zero phase shift. Thus the resultant feedback voltage V_f is 180° out of phase with the input signal V_{in}.

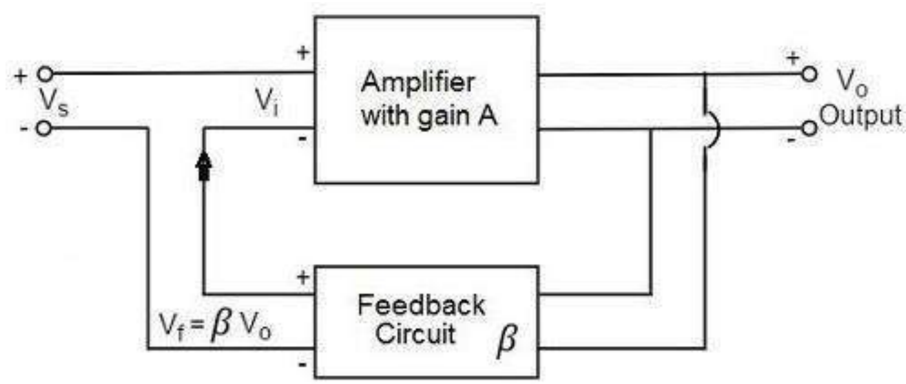
Though the gain of negative feedback amplifier is reduced, there are many advantages of negative feedback such as

- Stability of gain is improved
- Reduction in distortion
- Reduction in noise
- Increase in input impedance
- Decrease in output impedance
- Increase in the range of uniform application

It is because of these advantages negative feedback is frequently employed in amplifiers.

NEGATIVE FEEDBACK CIRCUIT:

A feedback amplifier generally consists of two parts. They are the amplifier and the feedback circuit. The feedback circuit usually consists of resistors. The concept of feedback amplifier can be understood from the following figure.



From the above figure, the gain of the amplifier is represented as A. the gain of the amplifier is the ratio of output voltage V_o to the input voltage V_i. the feedback network extracts a voltage $V_f = \beta\, V_o$ from the output V_o of the amplifier.

This voltage is added for positive feedback and subtracted for negative feedback, from the signal voltage V_s. Now,

$$V_i = V_s + V_f = V_s + \beta\, V_o$$

$$V_i = V_s - V_f = V_s + \beta\, V_o$$

The quantity $\beta = V_f/V_o$ is called as feedback ratio or feedback fraction.

Let us consider the case of negative feedback. The output V_o must be equal to the input voltage (V_s - βV_o) multiplied by the gain A of the amplifier.

Hence, $\quad (V_s - \beta\, V_o)\, A = V_o$

Or $\quad AV_s - A\beta V_o = V_o$

Or $\quad AV_s = V(1 + A\beta)$

Therefore, $\quad \frac{Vo}{V_s} = \frac{A}{1+A\beta}$

Let A_f be the overall gain (gain with the feedback) of the amplifier. This is defined as the ratio of output voltage V_o to the applied signal voltage V_s, i.e.

$$A_f = \frac{Output\ voltage}{Input\ signal\ Voltage} = \frac{V_o}{V_s}$$

So, from the above two equations, we can understand that.

The equation of gain of the feedback amplifier, with negative feedback is given by

$$A_f = \frac{A}{1 + A\beta}$$

The equation of gain of the feedback amplifier, with positive feedback is given by

$$A_f = \frac{A}{1 - A\beta}$$

These are the standard equations to calculate the gain of feedback amplifiers.

ADVANTAGES OF NEGATIVE FEEDBACK AMPLIFIER:

There are some advantages of negative feedback amplifier which are given below,

- The negative feedback reduces noise.
- It has highly stabilized gain.
- It can control step response of amplifier.
- It has less harmonic distortion.
- It has less amplitude distortion.
- It has less phase distortion.
- Input and output impedances can be modified as desired.
- It can increase or decrease output impedances.
- It has higher fidelity i.e. more linear operation.
- It has less frequency distortion.

There are some disadvantages of negative feedback amplifier which are given below,

- It Increase output resistance in case of current shunt and current series feedback amplifiers.
- Reduction in gain.

POWER AMPLIFIERS AND ITS CLASSIFICATION:

A power amplifier receives a signal from some pick-up transducer or other input source and provides a large version of the signal to some output device or to another amplifier stage.

On the basic of the amount of the output signal varying over one cycle of operation for a full cycle of input signal, large signal or power amplifiers are classified into five types. They are:-

i) CLASS A POWER AMPLIFIER:

For Class A Power amplifier the output signal varies for a full 360° of the cycle.

ii) CLASS B POWER AMPLIFIERS:

For Class B Power amplifier the output signal varies for 180° of the input signal.

iii) CLASS AB POWER AMPLIFIERS:

For class AB power amplifier, the output signal varies between 180° and 360°. It is neither a class A nor a class B amplifier.

iv) CLASS C POWER AMPLIFIERS:

For Class C Power amplifier the collector current flows for less than half cycle of the input signal.

v) CLASS D POWER AMPLIFIERS:

Class D Power amplifiers remain ON for a short interval and OFF for a long interval.

DIFFERENCE BETWEEN VOLTAGE AND POWER AMPLIFIER:

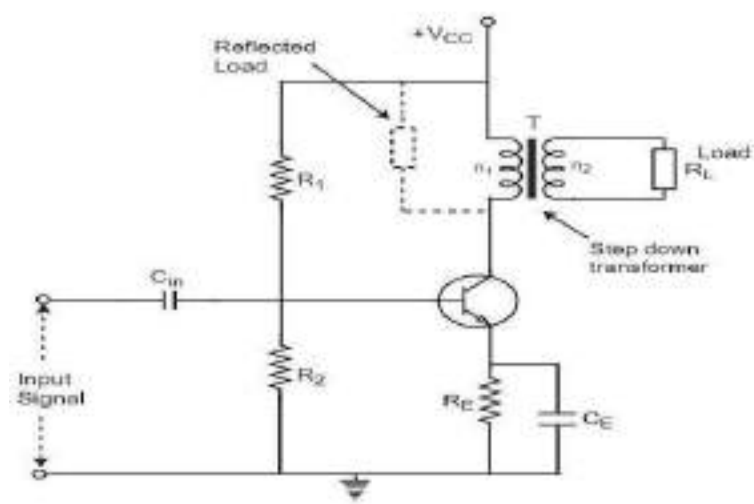
BASIS OF COMPARISON	VOLTAGE	POWER AMPLIFIER
Use	The voltage amplifier amplifies the voltage or increases the voltage level of a signal.	The power amplifier amplifies the power of a signal.
Functionality	The voltage amplifier can work with low magnitude signal.	The voltage amplifier can work with low magnitude signal.
Amplitude Of AC Signal	In voltage amplifier, the amplitude of input A.C signal is small.	In power amplifier the amplitude of input A.C signal is large.
Transistor	The transistor used in the voltage amplifier has a thin base because it does not handle large current.	The transistor used in the power amplifier has a thick base because it handles the very large current.
Output Impedance	The output impedance of the voltage amplifier is very high, about 12 kilo-ohm.	The output impedance of the power amplifier is very low, up to 200 ohm, so that it can deliver a high current.
Collector Current	The collector current of the voltage amplifier is very low up to 1 mA.	The collector current of the power amplifier is high greater than 100 mA.
Transistor Heat Dissipation	The transistor used can dissipate less heat produced during its operation.	The transistor used can dissipate more heat produced as compared to voltage amplifier during its operation.
Size Of Transistor	The physical size of transistor used is usually	The physical size of transistor used is usually large and is

	small and is known as low or medium power transistor.	known as power transistor.
Collector Load	In power amplifier, the collector load has low resistance, typically 5Ω to 20Ω.	In voltage amplifier, the collector load has high resistance, typically 4Ω to 10kΩ.
Coupling	RC coupling is used in voltage amplifier.	Transformer coupling is used in power amplifier.
Application	Voltage amplifier is used for small signal voltage.	Power amplifier is used for high voltage signals.
Current Gain	The current gain of the power amplifier is very high.	The current gain of the power amplifier is very high.

TRANSFORMER COUPLED CLASS A POWER AMPLIFIER:

The class A power amplifier as discussed in the previous chapter, is the circuit in which the output current flows for the entire cycle of the AC input supply. We also have learnt about the disadvantages it has such as low output power and efficiency. In order to minimize those effects, the transformer coupled class A power amplifier has been introduced.

The **construction of class A power amplifier** can be understood with the help of below figure. This is similar to the normal amplifier circuit but connected with a transformer in the collector load.



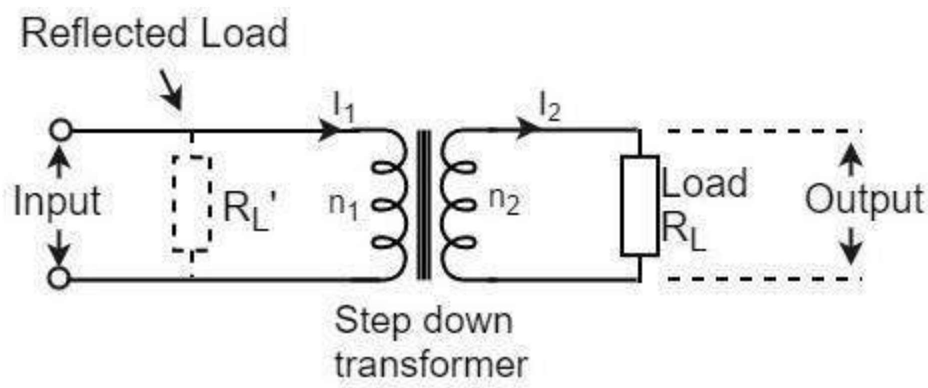
Here R_1 and R_2 provide potential divider arrangement. The resistor Re provides stabilization, C_e is the bypass capacitor and R_e to prevent a.c. voltage. The transformer used here is a step-down transformer.

The high impedance primary of the transformer is connected to the high impedance collector circuit. The low impedance secondary is connected to the load (generally loud speaker).

Transformer Action

The transformer used in the collector circuit is for impedance matching. R_L is the load connected in the secondary of a transformer. R_L' is the reflected load in the primary of the transformer.

The number of turns in the primary are n_1 and the secondary are n_2. Let V_1 and V_2 be the primary and secondary voltages and I_1 and I_2 be the primary and secondary currents respectively. The below figure shows the transformer clearly.



We know that

$$\frac{v_1}{v_2} = \frac{n_1}{n_2} \text{ and } \frac{I_1}{I_2} = \frac{n_1}{n_2}$$

Or

$$V_1 = \frac{n_1}{n_2} V_2 \text{ and } I_1 = \frac{n_1}{n_2} I_2$$

Hence

$$\frac{V_1}{I_1} = \left(\frac{n_1}{n_2}\right)^2 \frac{V_2}{I_2}$$

But $V_1/I_1 = R_L'$ = effective input resistance

And $V_2/I_2 = R_L$ = effective output resistance

Therefore,

$$R_L' = \left(\frac{n_1}{n_2}\right)^2 R_L = n^2 R_L$$

Where

$$n = \frac{\text{number of turns in primary}}{\text{number of turns in secondary}} = \frac{n_1}{n_2}$$

A power amplifier may be matched by taking proper turn ratio in step down transformer.

Circuit Operation

If the peak value of the collector current due to signal is equal to zero signal collector current, then the maximum a.c. power output is obtained. So, in order to achieve complete amplification, the operating point should lie at the center of the load line.

The operating point obviously varies when the signal is applied. The collector voltage varies in opposite phase to the collector current. The variation of collector voltage appears across the primary of the transformer.

Circuit Analysis

The power loss in the primary is assumed to be negligible, as its resistance is very small.

The input power under dc condition will be

$(P_{in})_c = (P_{tr})_{dc} = V_{CC} * (I_C)_Q$

Under maximum capacity of class A amplifier, voltage swings from $(V_{ce})_{max}$ to zero and current from $(I_c)_{max}$ to zero.

Hence

$$V_{rms} = \frac{1}{\sqrt{2}}\left[\frac{(V_{ce})ax-(V_{ce})min}{2}\right] = \frac{1}{\sqrt{2}}\left[\frac{(V_{ce})max}{2}\right] = \frac{2V_{CC}}{2\sqrt{2}} = \frac{V_{CC}}{\sqrt{2}}$$

$$I_{rms} = \frac{1}{\sqrt{2}}\left[\frac{(I_C)ax-(I_C)min}{2}\right] = \frac{1}{\sqrt{2}}\left[\frac{(I_C)max}{2}\right] = \frac{2(I_C)_Q}{2\sqrt{2}} = \frac{(I_C)_Q}{\sqrt{2}}$$

Therefore,

$$(P_o)_{ac} = V_{rms} * I_{rms} = \frac{V_{CC}}{\sqrt{2}} * \frac{(I_C)_Q}{\sqrt{2}} = \frac{V_{CC}*(I_C)_Q}{2}$$

Therefore,

$$\text{Collector Efficiency} = \frac{(Po)c}{(P_{tr})_{dc}}$$

Or,

$$(\eta)_{collector} = \frac{V_{cc} * (I_C)}{2 * V_{CC} * (I_C)_Q} = \frac{1}{2}$$

$$= \frac{1}{2} * 100 = 50\%$$

The efficiency of a class A power amplifier is nearly than 30% whereas it has got improved to 50% by using the transformer coupled class A power amplifier.

Advantages

The advantages of transformer coupled class A power amplifier are as follows.

- No loss of signal power in the base or collector resistors.
- Excellent impedance matching is achieved.
- Gain is high.

- DC isolation is provided.

Disadvantages

The disadvantages of transformer coupled class A power amplifier are as follows.

- Low frequency signals are less amplified comparatively.
- Hum noise is introduced by transformers.
- Transformers are bulky and costly.
- Poor frequency response.

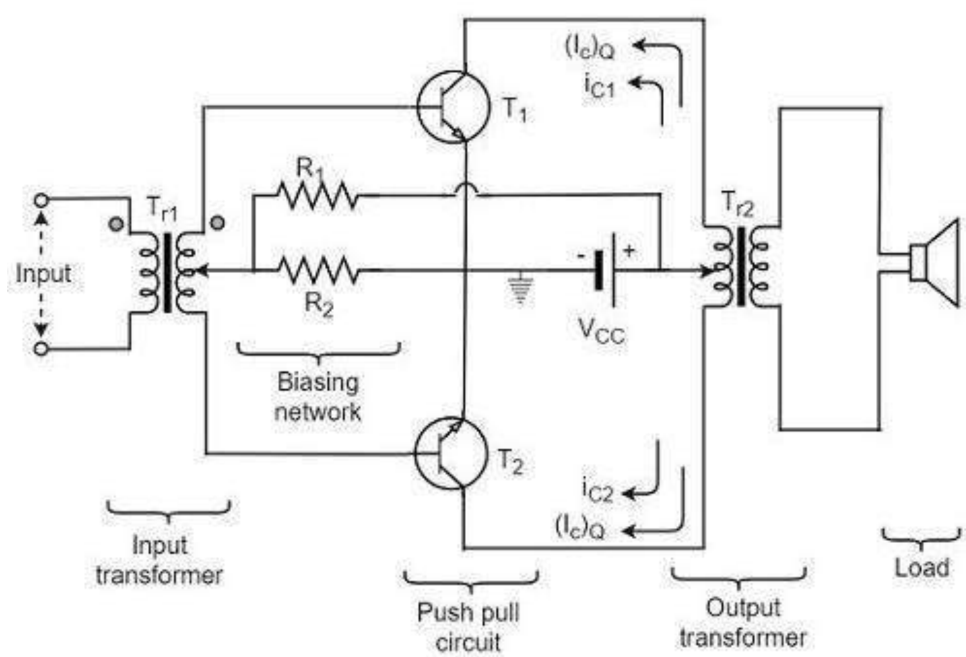
Applications

The applications of transformer coupled class A power amplifier are as follows.

- This circuit is where impedance matching is the main criterion.
- These are used as driver amplifiers and sometimes as output amplifiers.

CLASS A PUSH PULL AMPLIFIER:-

- ✓ Class A push pull amplifier uses two complementary transistors in the output stage with one transistor being an NPN or N-channel type while the other transistor is a PNP or P-channel (the complement) type connected in order to operate them like PUSH a transistor to ON and PULL another transistor to OFF at the same time.
- ✓ This push-pull configuration can be made in class A, class B, class C or class AB amplifiers.



Circuit Operation

The output is collected from the output transformer T_{r2}. The primary of this transformer T_{r2} has practically no dc component through it. The transistors T_1 and T_2 have their collectors connected to the primary of transformer T_{r2} so that their currents are equal in magnitude and flow in opposite directions through the primary of transformer T_{r2}.

When the a.c. input signal is applied, the base of transistor T_1 is more positive while the base of transistor T_2 is less positive. Hence the collector current i_{c1} of transistor T_1 increases while the collector current i_{c2} of transistor T_2 decreases. These currents flow in opposite directions in two halves of the primary of output transformer. Moreover, the flux produced by these currents will also be in opposite directions.

Hence, the voltage across the load will be induced voltage whose magnitude will be proportional to the difference of collector currents i.e.

$$(i_{c1}-i_{c2})$$

Similarly, for the negative input signal, the collector current i_{c2} will be more than i_{c1}. In this case, the voltage developed across the load will again be due to the difference

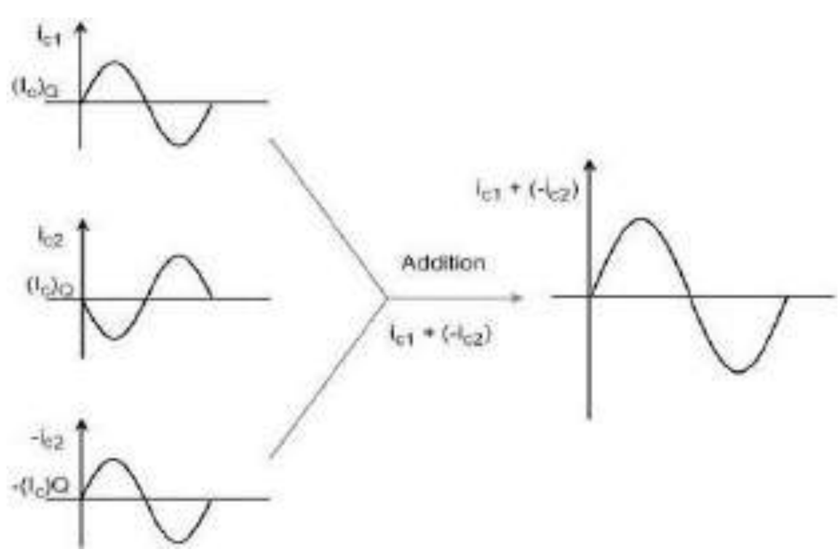
$$(i_{c1}-i_{c2})$$

$$\text{As } i_{c2}>i_{c1}$$

The polarity of voltage induced across load will be reversed.

$$i_{c1}-i_{c2}=i_{c1}+(-i_{c2})$$

To have a better understanding, let us consider the below figure.



The overall operation results in an a.c. voltage induced in the secondary of output transformer and hence a.c. power is delivered to that load.

Advantages

The advantages of class A Push-pull amplifier are as follows

- High a.c. output is obtained.
- The output is free from even harmonics.
- The effects of ripple voltages are balanced out. These are present in the power supply due to inadequate filtering.

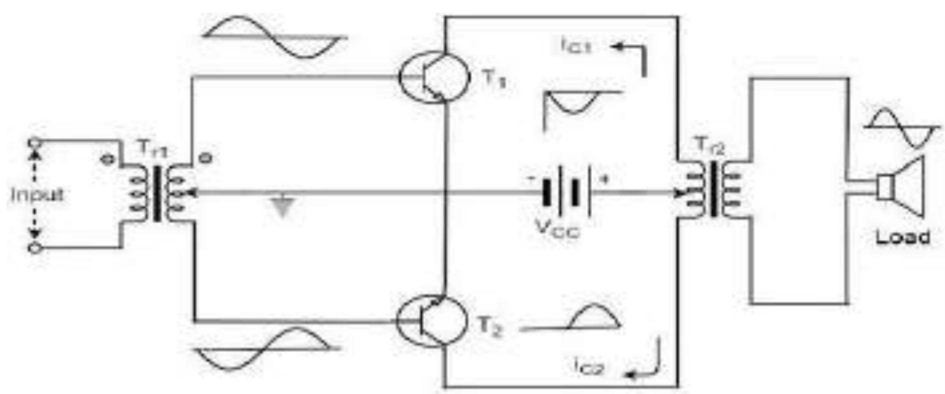
Disadvantages

The disadvantages of class A Push-pull amplifier are as follows

- The transistors are to be identical, to produce equal amplification.
- Center-tapping is required for the transformers.
- The transformers are bulky and costly.

CLASS B PUSH PULL AMPLIFIER:-

Though the efficiency of class B power amplifier is higher than class A, as only one half cycle of the input is used, the distortion is high. Also, the input power is not completely utilized. In order to compensate these problems, the push-pull configuration is introduced in class B amplifier.

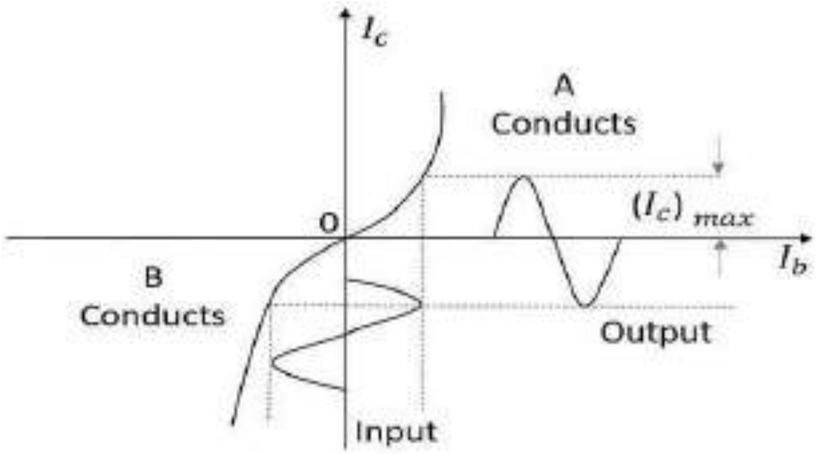


The circuit operation of class B push pull amplifier is given below.

Operation

The circuit of class B push-pull amplifier shown in the above figure clears that both the transformers are center-tapped. When no signal is applied at the input, the transistors T_1 and T_2 are in cut off condition and hence no collector currents flow. As no current is drawn from V_{CC}, no power is wasted.

When input signal is given, it is applied to the input transformer T_{r1} which splits the signal into two signals that are 180^o out of phase with each other. These two signals are given to the two identical transistors T_1 and T_2. For the positive half cycle, the base of the transistor T_1 becomes positive and collector current flows. At the same time, the transistor T_2 has negative half cycle, which throws the transistor T_2 into cutoff condition and hence no collector current flows. The waveform is produced as shown in the following figure.



For the next half cycle, the transistor T_1 gets into cut off condition and the transistor T_2 gets into conduction, to contribute the output. Hence for both the cycles, each transistor conducts alternately. The output transformer T_{r3} serves to join the two currents producing an almost undistorted output waveform.

Power Efficiency of Class B Push-Pull Amplifier

The current in each transistor is the average value of half sine loop.

For half sine loop, I_{dc} is given by

$$I_{dc} = \frac{(I_C)max}{\pi}$$

Therefore, $(P_{in})_c = 2 \text{ X } [\frac{(IC)max}{\pi} \text{ X } V_{CC}]$

Here factor 2 is introduced as there are two transistors in push pull amplifier.

R.M.S value of collector current = ${(I_C)ax}/{\sqrt{2}}$

R.M.S value of output voltage = ${V_{CC}}/{\sqrt{2}}$

Under ideal conditions of maximum power

Therefore, $(P_O)_c = \frac{(IC)max}{\sqrt{2}} \text{ X } \frac{VCC}{\sqrt{2}} = \frac{(IC)max \text{ X } VCC}{2}$

Now overall maximum efficiency

$\eta_{overall} = \frac{(P_O)_c}{(P_{in})_{dc}}$

$= \frac{(I_C)max \text{ X } V_{CC}}{2} \text{ X } \frac{\pi}{2(I_C)_{ax} \text{ X } V_{CC}}$

$= \frac{\pi}{4} = 0.785 = 78.5\%$

The collector efficiency would be the same.

Hence the class B push pull amplifier improves the efficiency than the class A push pull amplifier.

OSCILLATORS

An oscillator is a circuit which produces a continuous, repeated, alternating waveform without any input. Oscillators basically convert unidirectional current flow from a DC source into an alternating waveform which is of the desired frequency, as decided by its circuit components.

The following points may be noted:

1. A transistor amplifier with proper positive feedback will work as an oscillator.
2. The circuit needs only a quick trigger signal to start the oscillations. Once the oscillations have started, no external signal source is needed.
3. In order to get continuous Undamped output from the circuit, the following condition must be met:

$$m_v A_v = 1$$

Where m_v = voltage gain of amplifier without feedback
A_v = feedback fraction

This relation is called Barkhausen criterion.

Barkhausen Criterion

Barkhausen criterion is that in order to produce continuous Undamped oscillations at the output of an amplifier, the positive feedback should be such that :

$$m_v A_v = 1$$

Once this condition is set in the positive feedback amplifier, continuous undamped oscillations can be obtained at the output.

Different types of Transistor Oscillators

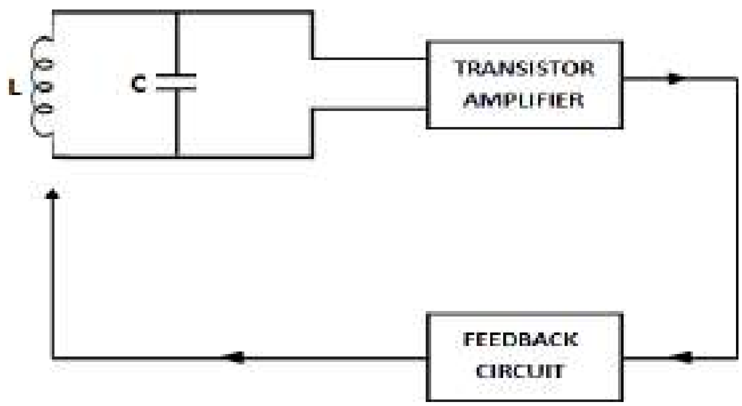
All oscillators under different names have similar function i.e. they produce continuous undamped output. However, the major difference between these oscillators lies in the method by which energy is supplied to the tank circuit to meet the losses.

The following are the transistor oscillators commonly used at various places in electronic circuit:

1. Tuned collector oscillator
2. Colpitt Oscillator
3. Hartley Oscillator
4. Phase shift Oscillator
5. Wien Bridge Oscillator
6. Crystal Oscillator

Essentials of Transistor Oscillator

the block diagram of an oscillator is shown below. Its essential components are:



(i) Tank Circuit:

It consists of inductance (L) connected in parallel with capacitor (C). The frequency of oscillations in the circuit depends upon the values L and C.

(ii) Transistor amplifier:

The transistor amplifier receives d.c. power from the battery and changes it into a.c. power for supplying to the tank circuit. The oscillations occurring in the tank circuit are applied to the input of the transistor amplifier. Because of the amplifying nature of the transistor, we get increased output of these oscillations. This amplified output of oscillations is due to the d.c. power supplied by the battery. The output of the transistor can be supplied to the tank circuit to meet the losses.

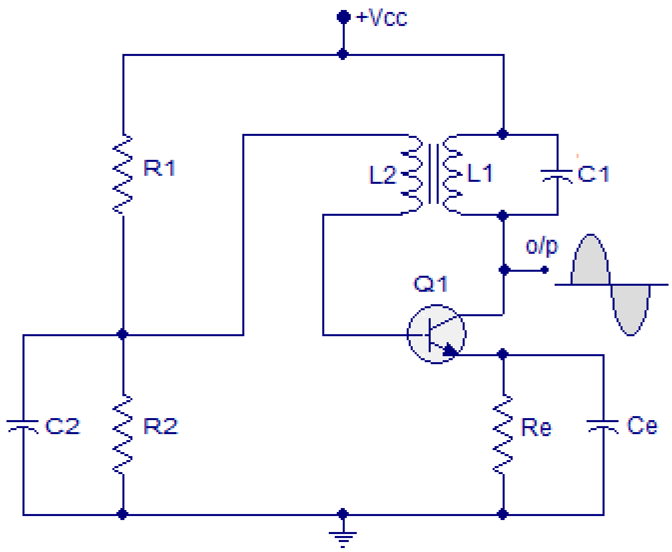
(iii) Feedback circuit:

The feedback circuit supplies a part of collector energy to the tank circuit in correct phase to aid the oscillations i.e. it provides positive feedback.

PRINCIPLE OF OPERATION OF TUNED COLLECTOR, HARTLEY, COLPITT,PHASE SHIFT, WEIN- BRIDGE OSCILLATOR

TUNED COLLECTOR OSCILLATOR: -

- ✓ Tuned collector oscillator is the simplest and the basic type of LC oscillators.



Tuned collector oscillator

Circuit Operation:

When V_{CC} is applied to the circuit, collector current starts increasing and charges the capacitor C_1.

When this capacitor is fully charged, it discharges through coil L_1, setting up oscillations of frequency determined by exp.(i).

These oscillations induce some voltage in coil L_2 by mutual induction.

The frequency of voltage in coil L_2 is the same as that of tank circuit but its magnitude depends upon the number of turns of L_2 and coupling between L_1 and L_2.

The voltage across L_2 is applied between base and emitter and appears in the amplified form in the collector circuit, thus overcoming the losses occurring in the tank circuit.

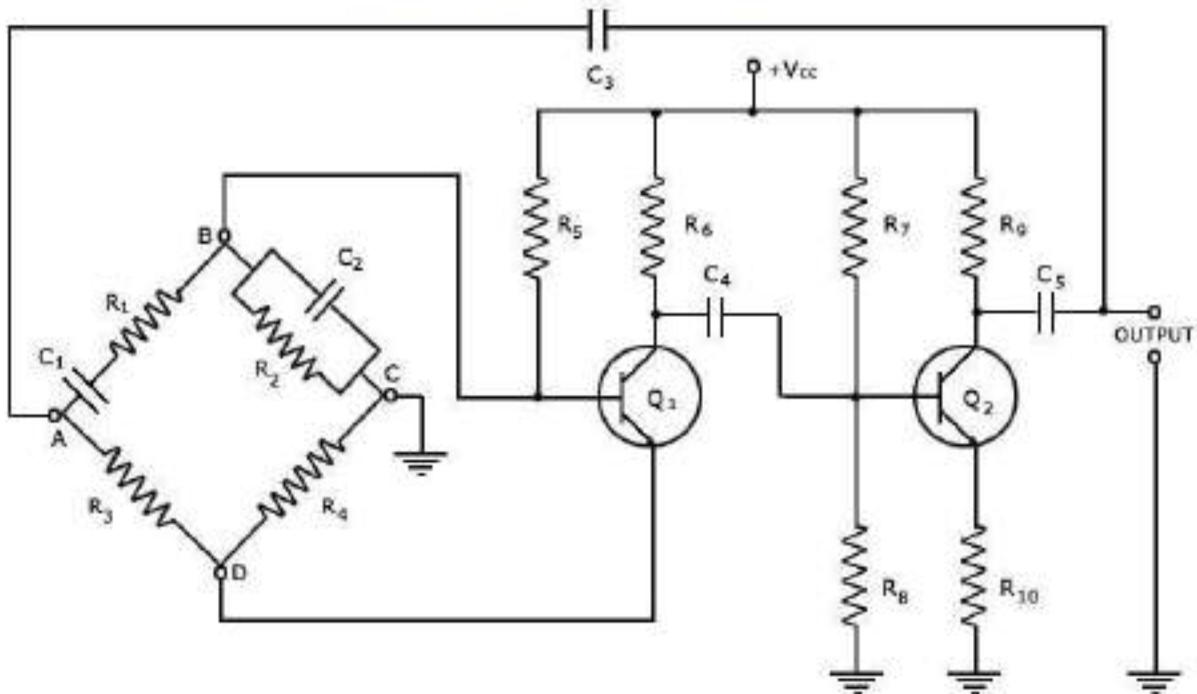
The number of turns of L_2 and coupling between L_1 and L_2 are so adjusted that oscillations across L_2 are amplified to a level just sufficient to supply losses to the tank circuit.

A phase shift of 180^o is created between the voltage of L_1 and L_2 due to transformer action. A further phase shift of 180^o takes place between base-emitter and collector circuit due to transistor properties. As a result, the energy feedback to the tank circuit is in phase with the generated oscillations.

WEIN BRIDGE OSCILLATOR:-

- ✓ It is essentially a two-stage amplifier with an R-C bridge circuit.

Wien Bridge Oscillator Circuit



Wien Bridge Oscillator – Working

- ✓ The circuit is set in oscillation by any random change in base current of transistor Q_1 that may be due to noise inherent in the transistor or variation in voltage of dc supply.
- ✓ This variation in base current is amplified in collector circuit of transistor Q_1 but with a phase-shift of 180°.
- ✓ The output of transistor Q_1 is fed to the base of second transistor Q_2 through capacitor C_4.
- ✓ Now a still further amplified and twice phase-reversed signal appears at the collector of the transistor Q_2.
- ✓ Having been inverted twice, the output signal will be in phase with the signal input to the base of transistor Q_1.
- ✓ The continuous frequency variation in this oscillator can be done by varying the two capacitors C_1 and C_2 simultaneously.
- ✓ These capacitors are variable air-gang capacitors.
- ✓ The advantages and disadvantages of Wien bridge oscillators are given below:

Advantages

1. Provides a stable low distortion sinusoidal output over a wide range of frequency.
2. The frequency range can be selected simply by using decade resistance boxes.
3. The frequency of oscillation can be easily varied by varying capacitances C_1 and C_2 simultaneously. The overall gain is high because of two transistors.

Disadvantages

1. The circuit needs two transistors and a large number of other components.

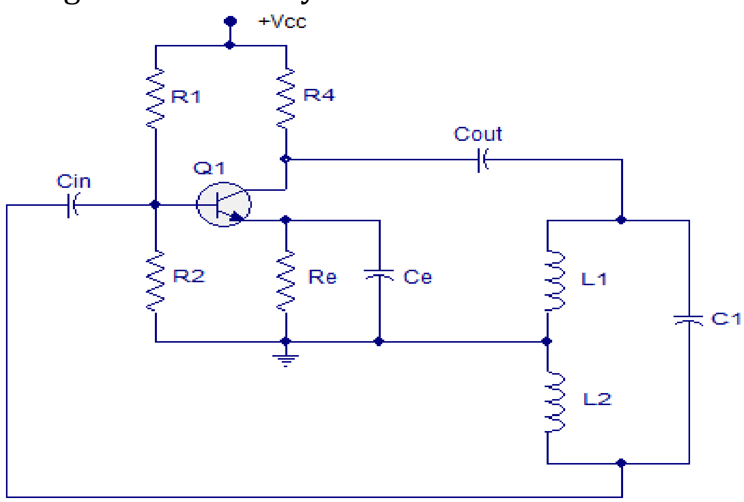
2. The maximum frequency output is limited because of amplitude and the phase-shift characteristics of amplifier.

HARTLEY OSCILLATOR:-

- ✓ The Hartley oscillator is an electronic oscillator circuit in which the oscillation frequency is determined by the tuned circuit consisting of capacitors and inductors, that is, an LC oscillator.

Hartley Oscillator Circuit and Working

- ✓ The circuit diagram of a Hartley oscillator is shown in the below figure.



Hartley oscillator

- ✓ When the power supply is switched ON the transistor starts conducting and the collector current increases.
- ✓ As a result the capacitor C1 starts charging and when the capacitor C1 is fully charged it starts discharging through coil L1.
- ✓ This charging and discharging creates a series of damped oscillations in the tank circuit and it is the key.
- ✓ The oscillations produced in the tank circuit is coupled (fed back) to the base of Q1 and it appears in the amplified form across the collector and emitter of the transistor.
- ✓ The output voltage of the transistor (voltage across collector and emitter) will be in phase with the voltage across inductor L1.
- ✓ Since the junction of two inductors is grounded, the voltage across L2 will be 180° out of phase to that of the voltage across L1.
- ✓ The voltage across L2 is actually fed back to the base of Q1.
- ✓ From this we can see that, the feedback voltage is 180° out of phase with the transistor and also the transistor itself will create another 180° phase difference.
- ✓ So the total phase difference between input and output is 360° and it is very important condition for creating sustained oscillations.

- ✓ The frequency of oscillations in this circuit is

  $f_o = 1/ (2\pi \sqrt{(L_{eq} C)})$

- ✓ Where Leq is the total inductance of coils in the tank circuit is given as

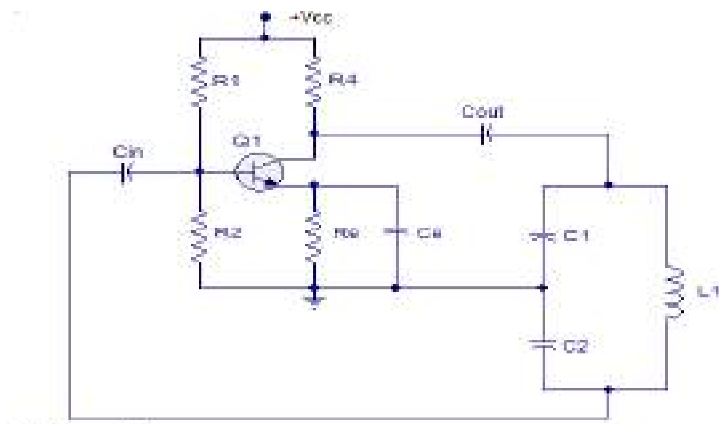
  $L_{eq} = L1 + L2 + 2M$

- ✓ For a practical circuit, if L1 = L2 = L and the mutual inductance are neglected, then the frequency of oscillations can be simplified as

  $f_o = 1/ (2\pi \sqrt{(2 L C)})$

COLPITT OSCILLATOR:-

- ✓ Colpitt Oscillator is an electronic oscillator which uses an inductor and capacitors to form an LC oscillator circuit.
- ✓ Colpitt oscillator was invented by American scientist Edwin Colpitt in 1918.
- ✓ The circuit diagram of a typical Colpitt oscillator using transistor is shown in the figure below.



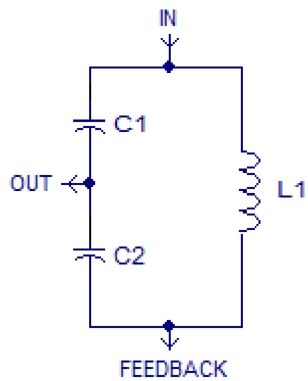
Colpitts oscillator

- ✓ When power supply is switched ON, capacitors C1 and C2 starts charging.
- ✓ When they are fully charged they start discharging through the inductor L1.
- ✓ When the capacitors are fully discharged, the electrostatic energy stored in the capacitors gets transferred to the inductor as magnetic flux.
- ✓ The inductor starts discharging and capacitors get charged again.
- ✓ This transfer of energy back and forth between capacitors and inductor is the basis of oscillation.
- ✓ Voltage across C2 is phase opposite to that of the voltage across the C1 and it is the voltage across C2 that is fed back to the transistor.
- ✓ The feedback signal at the base of transistor appears in the amplified form across the collector and emitter of the transistor.

- ✓ The energy lost in the tank circuit is compensated by the transistor and the oscillations are sustained.
- ✓ The tank circuit produces 180° phase shift and the transistor itself produces another 180° phase shift.
- ✓ That means the input and output are in phase and it is a necessary condition of positive feedback for maintaining sustained oscillations.
- ✓ The frequency of oscillations of the Colpitt oscillator can be determined using the equation below.

$$F = \frac{1}{2\pi\sqrt{LC}}$$

Where L is the inductance of the inductor in the tank circuit and
C is the effective capacitance of the capacitors in the tank circuit.



- ✓ If C1 and C2 are the individual capacitance, then the effective capacitance of the serial combination C= (C1C2)/ (C1+C2).
- ✓ By using ganged variable capacitors in place of C1 and C2, the Colpitt oscillator can be made variable.

Advantages of Colpitt oscillator

- ✓ Main advantage of Colpitt oscillator over Hartley oscillator is the improved performance in the high frequency region.
- ✓ This is because the capacitors provide a low reactance path for the high frequency signals and thus the output signals in the high frequency domain will be more sinusoidal.
- ✓ Due to the excellent performance in the high frequency region, the Colpitt oscillator can be even used in microwave applications.

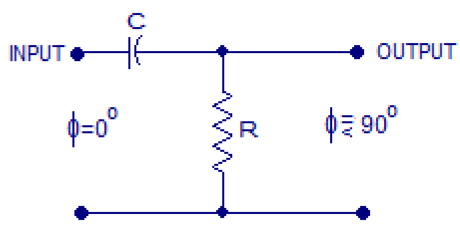
PHASE SHIFT OSCILLATOR:

- ✓ RC phase shift oscillator or simply RC oscillator is a type of oscillator where a simple RC network (resistor-capacitor) network is used for giving the required phase shift to the feedback signal.

- In LC oscillators like Hartley oscillator and Colpitt oscillator an LC network (inductor- capacitor network) is used for providing the necessary positive feedback.
- The main feature of an RC phase shift oscillator is the excellent frequency stability.
- The RC oscillator gives a pure sine wave output on a wide range of loads.

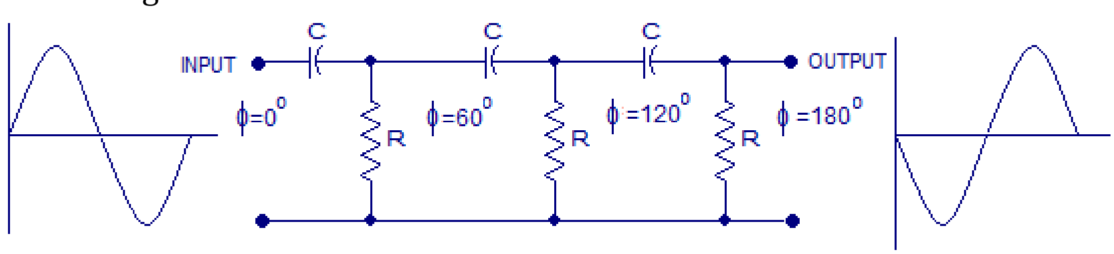
RC phase shift network.

- RC phase shift network is a simple resistor capacitor network that can be used to give a desired phase shift to a signal.
- The circuit diagram of a simple single stage RC network is shown in the figure below.



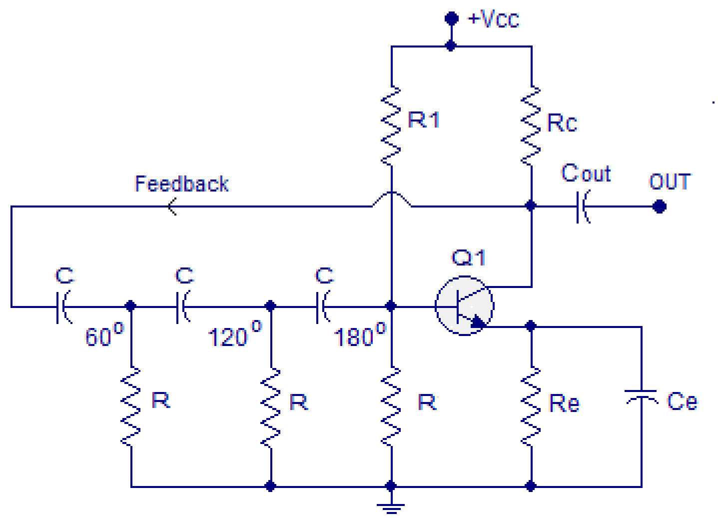
Single stage RC phase shift network

- Theoretically in a simple RC circuit, the output voltage will lead the input voltage by a phase angle Φ =90°.
- But in practical case the phase angle will be something below 90° just because it is impossible to get a purely ideal capacitor.
- Phase shift of a practical RC network depends on the value of the capacitor, resistor and the operating frequency.
- Let F be the operating frequency,
  R be the resistance and C be the capacitance.
- Then the capacitive reactance Xc to the frequency F can be given by the equation $Xc = 1 / (2\pi FC)$
- The effective impedance of the circuit can be given by the equation $Z = \sqrt{(R^2 + Xc^2)}$
- The phase angle of the RC network can be derived as $\Phi = \tan^{-1} (Xc/R)$.
- Just by making an RC network with phase shift equal to 60° and cascading three of them together the desired phase shift of 180° can be attained.
- This 180° phase shift by the RC network plus the 180° phase shift made by the transistor gives a total phase shift of 360° between the input and output which is the necessary condition for maintaining sustained oscillations.
- The circuit diagram of a three stage RC network producing a phase shift of 180° is shown in the figure below.



Three stage RC phase shift network

- ✓ Connecting such a three stage RC phase shift network between the input and output of a common emitter transistor amplifier will result in a transistor based RC phase shift oscillator.
- ✓ The circuit diagram is shown below.



RC Phase shift oscillator

- ✓ In the circuit diagram resistor R1 and the resistor R (close to the base of Q1 in the diagram) gives a voltage divider bias to the transistor Q1.
- ✓ Resistor Rc limits the collector current while Re is meant for thermal stability.
- ✓ Ce is the emitter by-pass capacitor and Cout is the output DC decoupling capacitor.
- ✓ By using more than three RC phase shift stages (like 4 x 45°) the frequency stability of the oscillator can be further improved.
- ✓ The frequency of the transistor RC phase shift oscillator can be expressed by the equation:

$$F = \frac{1}{2\pi RC\sqrt{2N}}$$

Where F is the frequency, R is the resistance, C is the capacitance and N is the number of RC phase shift stages.

- ✓ The RC phase shift oscillator can be made variable by making the resistors or capacitors variable.
- ✓ The common approach is to leave the resistors untouched while the three capacitors are replaced by a triple gang variable capacitor.

UNIT-7: FIELD EFFECT TRANSISTOR

FET (FIELD EFFECT TRANSISTOR):

Field effect transistor or FET is a transistor, where output current is controlled by electric field. FET sometimes, is called uni-polar transistor as it involves single carrier type operation. The basic type of FET transistor is completely different from BJT transistor basics. FET is three terminal semiconductor devices, with source, drain and gate terminals.

The charge carries are electrons or holes, which flow from the source to drain though an active channel. This flow of electrons from source to drain is controlled by voltage applied across the gate and source terminals.

TYPES OF FET TRANSISTOR

FETs are of two types- JFETs or MOSFETs.

Junction FET

The Junction FET transistor is a type of field effect transistor that can be used as an electrically controlled switch. The electric energy flows through an active channel between sources to drain terminals. By applying a reverse bias voltage to gate terminal, the channel is strained so the electric current is switched off completely.

The junction FET transistor is available in two polarities which are;

1. N-Channel JFET
2. P-Channel JFET

MOSFET Transistor

MOSFET transistor as its name suggests is a p type (n type) semiconductor bar (with two heavily doped n type regions diffused into it) with a metal oxide layer deposited on its surface and holes taken out of the layer to form source and drain terminals. A metal layer is deposited on the oxide layer to form the gate terminal. One of the basic applications of field effect transistor is using a MOSFET as a switch.
This type of FET transistor has three terminals, which are source, drain, and gate. The voltage applied to the gate terminal controls the flow of current from source to drain. The presence of an insulating layer of metal oxide results in the device having high input impedance.

MOSFET is of two types. They are:

1. Enhancement type MOSFET
2. Depletion type MOSFET

ADVANTAGES OF FET OVER BJT

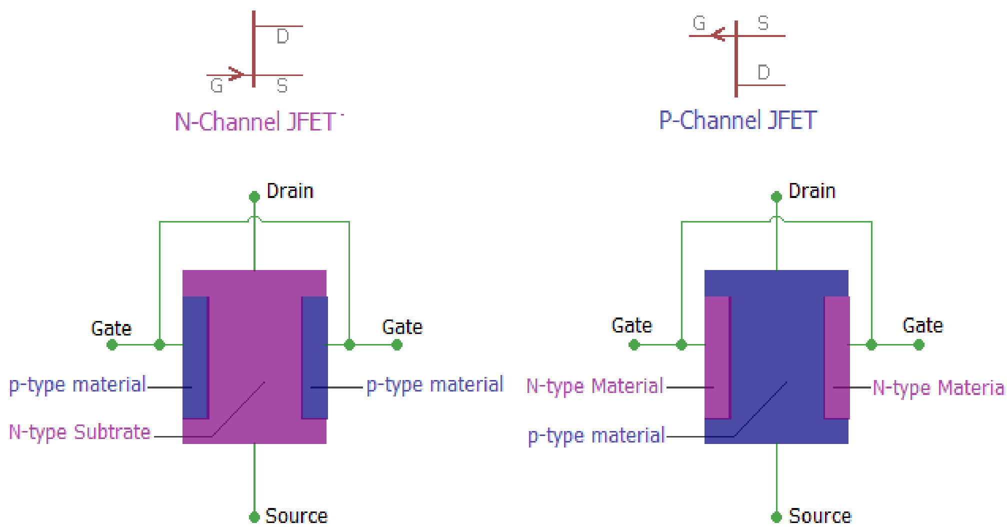
There are many advantages of FET's over BJT's here I'm providing few of them as follows

1. BJT'S are highly noisy devices than the FET's.
2. BJT's causes much and more loading effect than that of the FET's. Which is the major disadvantage of using BJT in circuit where loading effect should be avoided.

3. FET's are faster switching devices than that of the BJT. Because FET is majority carrier device.
4. FET's are smaller than BJT, thus FET takes lesser area on semiconductor die and they have higher input impedance than that of the BJT.
5. BJT's VI characteristics are linear while FET's have non-linear characteristics. Thus it's the advantages of using the BJT in its linear region for amplification process rather than using the FET.
6. Thermal stability of FET's are higher than the BJT's, because FET'S are unipolar devices that is current flow in FET depends upon the majority carriers only.

PRINCIPLE OPERATION OF JFET:

Construction of JFET



In the above image, we can see the basic construction of a JFET. The N-Channel JFET consists of P-type material in N-type substrate whereas N-type materials are used in the p-type substrate to form a P channel JFET.

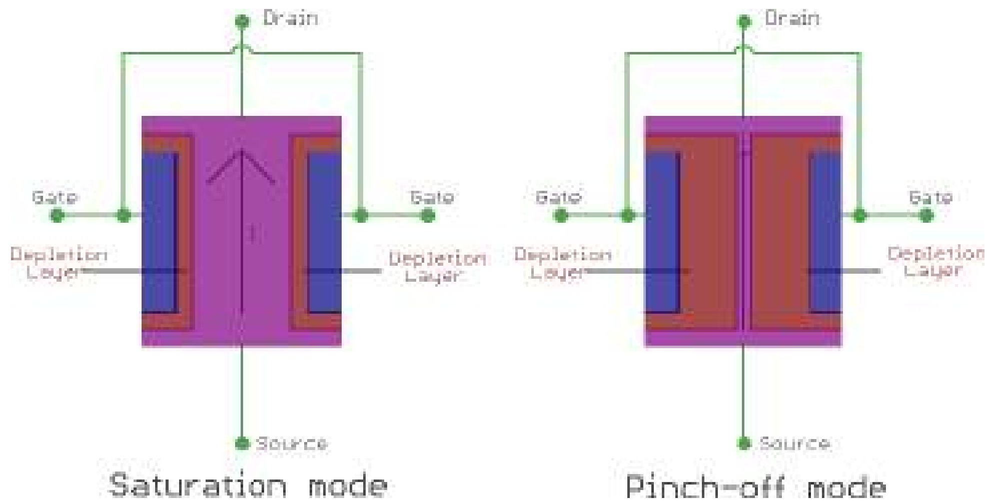
JFET is constructed using the long channel of semiconductor material. Depending on the construction process, if the JFET contains a great number of positive charge carriers (refers as holes) is a P-type JFET, and if it has a large number of negative charge carriers (refers as electrons) is called N-type JFET.

In the long channel of semiconductor material, Ohmic contacts at each end are created to form the Source and Drain connections. A P-N junction is formed in one or both side of the channel.

Working of JFET

When there is no voltage across gate and source, the channel becomes a smooth path which is wide open for electrons to flow. But the reverse thing happens when a voltage is applied between gate and source in reverse polarity that makes the P-N junction reversed biased and makes the channel narrower by increasing the depletion layer and could put the JFET in cut-off or pinch off region.

In the below image we can see the saturation mode and pinch off mode and we will be able to understand the depletion layer became wider and the current flow becomes less.



If we want to switch off a JFET we need to provide a negative gate to source voltage denoted as V_{GS} for an N-type JFET. For a P-type JFET, we need to provide positive V_{GS}.

JFET only works in the depletion mode, whereas MOSFETs have depletion mode and enhancement mode.

FET PARAMETERS:

During purchasing a JFET for a particular application we need to check the specifications of the device. These specifications are provided by manufacturers. The followings are the parameters used to specify a JFET and these are

- Gate Cut Off Voltage ($V_{GS(off)}$)
- Shorted Gate Drain Current (I_{DSS})
- Trans-conductance (g_{mo})
- Dynamic Output Resistance (r_d)
- Amplification Factor (μ)

Gate Cut Off Voltage

At a fixed drain voltage, the drain current (I_D) of a JFET depends on the gate to source voltage (V_{GS}). If the gate to source voltage decreases from zero in n channel JFET, the drain current also gets decreased accordingly. The relation between gate to source voltage and drain current is given below.

$$I_D = I_{DSS}\left[1 - \frac{V_{GS}}{V_{GS(off)}}\right]^2$$

When the gate terminal is grounded (V_{GS} = 0) and positive drain to source voltage (V_{DS}) is being increased slowly in case of n channel JFET, the drain current gets increased linearly. But after pinch-off voltage (V_p), the drain current would not be increased further and gets a constant value. This is the maximum drain current that can flow through the channel when the gate terminal is in ground potential. This current is fixed for a JFET and this is called shorted gated drain current and generally denoted by I_{DSS}.

Dynamic Output Resistance

This is the ratio of change of drain to source voltage (δV_{DS}) to the change of drain current (δI_D) at a constant gate to source voltage (V_{GS} = Constant). The ratio is denoted as r_d.

$$r_d = \frac{\delta V_{DS}}{\delta I_D} \quad at\ constant\ \ V_{GS}$$

Amplification Factor

The amplification factor is defined as the ratio of change of drain voltage (δV_{DS}) to change of gate voltage (δV_{GS}) at a constant drain current (I_D = Constant).

$$\mu = \frac{\delta V_{DS}}{\delta V_{GS}} \quad at\ constant\ \ I_D$$

There is a relation between trans-conductance (g_m) and dynamic output resistance (r_d) and that can be established in the following way.

$$\mu = \frac{\delta V_{DS}}{\delta V_{GS}} = \frac{\delta V_{DS}}{\delta I_D} \times \frac{\delta I_D}{\delta V_{GS}}$$

$$\Rightarrow \mu = r_d \times g_m$$

Trans-conductance

Trans-conductance is the ratio of change in drain current (δI_D) to change in the gate to source voltage (δV_{GS}) at a constant drain to source voltage (V_{DS} = Constant).

$$g_m = \frac{\delta I_D}{\delta V_{GS}} \quad at\ constant\ \ V_{DS}$$

This value is maximum at V_{GS} = 0. This is denoted by g_{mo}. This maximum value (g_{mo}) is specified in a JFET data sheet. The trans-conductance at any other value of gate to source voltage (g_m) can be determined as follows. The expression of drain current (I_D) is

$$I_D = I_{DSS}\left[1 - \frac{V_{GS}}{V_{GS(off)}}\right]^2$$

By partial differentiating the expression of drain current (I_D) in respect of gate to source voltage (V_{GS})

$$g_m = \frac{\delta I_D}{\delta V_{GS}} = \frac{2I_{DSS}}{V_{GS(off)}}\left[1 - \frac{V_{GS}}{V_{GS(off)}}\right]$$

At $V_{GS} = 0$, the trans-conductance gets its maximum value and that is

$$g_{mo} = \frac{2I_{DSS}}{V_{GS(off)}}$$

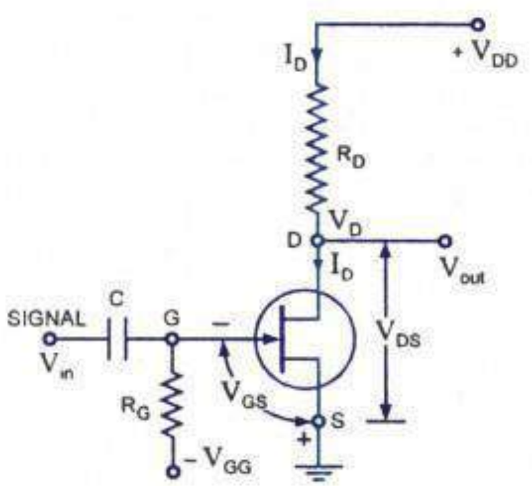
Therefore, we can write,

$$g_m = g_{mo}\left[1 - \frac{V_{GS}}{V_{GS(off)}}\right]$$

BIASING OF FET:

MOSFET bias circuits are similar to those used for JFETs. Various FET biasing circuits are discussed below:

<u>Fixed Bias.</u>



Fixed Biasing Circuit For JFET

DC bias of a FET device needs setting of gate-source voltage V_{GS} to give desired drain current I_D . For a JFET drain current is limited by the saturation current I_{DS}. Since the FET has such a high input impedance that no gate current flows and the dc voltage of the gate set by a voltage divider or a fixed battery voltage is not affected or loaded by the FET.

Fixed dc bias is obtained using a battery V_{QG}. This battery ensures that the gate is always negative with respect to source and no current flows through resistor R_G and gate terminal that is I_G =0. The battery provides a voltage V_{GS} to bias the N-channel JFET, but no resulting current is drawn from the battery V_{GG}. Resistor R_G is included to allow any ac signal applied through capacitor C to develop across R_G. While any ac signal will develop across R_G, the dc voltage drop across R_G is equal to I_G R_G i.e. 0 volt.

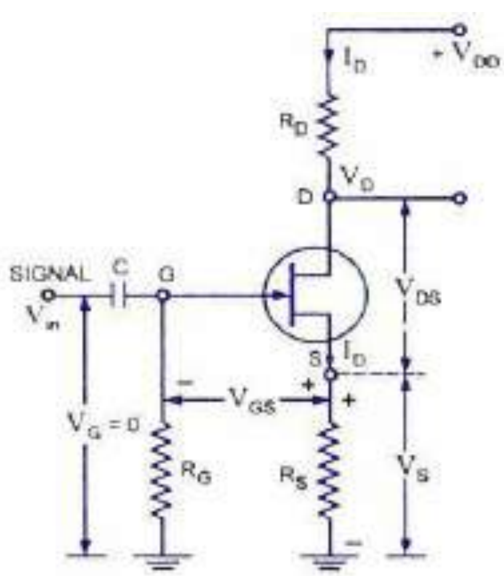
The gate-source voltage V_{GS} is then

$$V_{GS} = -V_G - V_s = -V_{GG} - 0 = -V_{GG}$$

The drain -source current I_D is then fixed by the gate-source voltage as determined by equation.

This current then causes a voltage drop across the drain resistor R_D and is given as $V_{RD} = I_D R_D$ and output voltage, $V_{out} = V_{DD} - I_D R_D$

<u>Self-Bias.</u>



Self-Bias Circuit For N-Channel JFET

This is the most common method for biasing a JFET. Self-bias circuit for N-channel JFET is shown in figure.

Since no gate current flows through the reverse-biased gate-source, the gate current $I_G = 0$ and, therefore, $v_G = i_G R_G = 0$

With a drain current I_D the voltage at the S is

$$V_s = I_D R_s$$

The gate-source voltage is then

$$V_{Gs} = V_G - V_s = 0 - I_D R_s = -I_D R_s$$

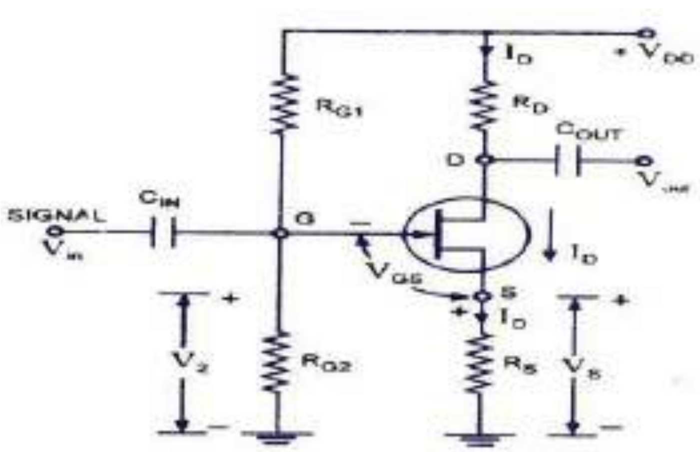
So voltage drop across resistance R_s provides the biasing voltage V_{Gg} and no external source is required for biasing and this is the reason that it is called self-biasing.

The operating point (that is zero signal I_D and V_{DS}) can easily be determined from equation and equation given below:

$$V_{DS} = V_{DD} - I_D (R_D + R_S)$$

Thus dc conditions of JFET amplifier are fully specified. Self-biasing of a JFET stabilizes its quiescent operating point against any change in its parameters like trans-conductance. Let the given JFET be replaced by another JFET having the double conductance then drain current will also try to be double but since any increase in voltage drop across R_s, therefore, gate-source voltage, V_{GS} becomes more negative and thus increase in drain current is reduced.

<u>Potential-Divider Biasing.</u>



Potential-Divider Bias Circuit
For N-Channel JFET

A slightly modified form of dc bias is provided by the circuit shown in figure. The resistors R_{Gl} and R_{G2} form a potential divider across drain supply V_{DD}. The voltage V_2 across R_{G2} provides the necessary bias. The additional gate resistor R_{Gl} from gate to supply voltage facilitates in larger adjustment of the dc bias point and permits use of larger valued R_S.

The gate is reverse biased so that $I_G = 0$ and gate voltage

$V_G = V_2 = (V_{DD}/R_{G1} + R_{G2}) * R_{G2}$

And

$V_{GS} = v_G - v_s = V_G - I_D R_s$

The circuit is so designed that $I_D R_s$ is greater than VG so that VGS is negative. This provides correct bias voltage.

The operating point can be determined as

$I_D = (V_2 - V_{GS}) / R_S$

And

$V_{DS} = V_{DD} - I_D (R_D + R_S)$

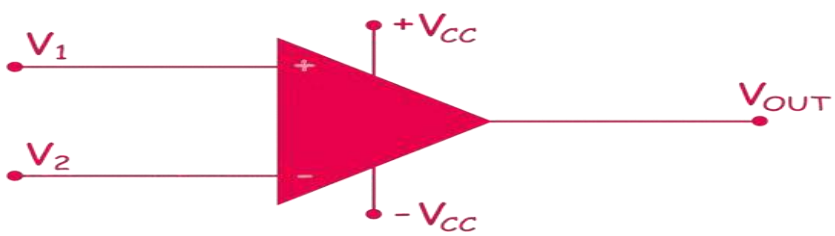
UNIT-8: OPERATIONAL AMPLIFIER

GENERAL CIRCUIT OF OP-AMP:

Op amp is basically a multistage amplifier in which a number of amplifier stages are interconnected to each other in a very complicated manner. Its internal circuit consists of many transistors, FETs and resistors. All this occupies a very little space.

So, it is packed in a small package and is available in the Integrated Circuit (IC) form. The term **Op Amp** is used to denote an amplifier which can be configured to perform various operations like amplification, subtraction, differentiation, addition, integration etc. An example is the very popular IC 741.

The symbol and its actual appearance in the IC form is shown below. The symbol appears as an arrowhead which signifies that the signal is flowing from output to input.



Input and Output Terminals of an Operational Amplifier

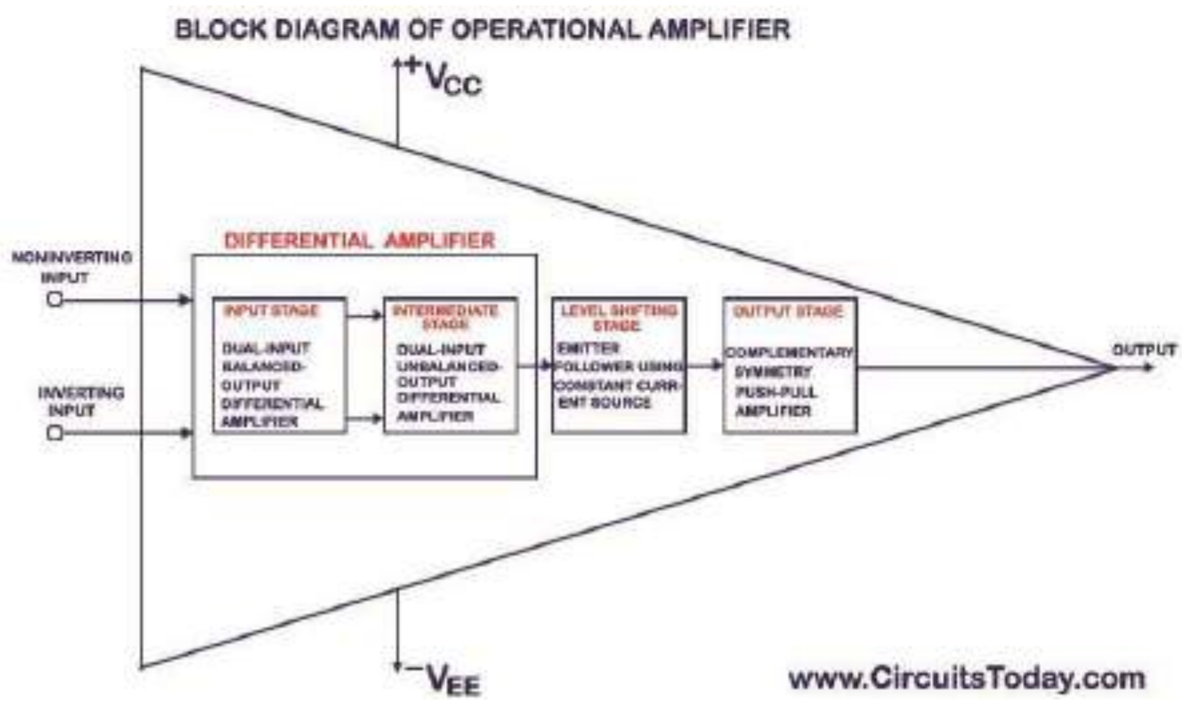
An op-amp has two input terminals and one output terminal. The op-amp also has two voltage supply terminals as seen above. Two input terminals form the differential input. We call the terminal, marked with negative (-) sign as the inverting terminal and the terminal marked with positive (+) sign as the non-inverting terminal of the **operational amplifier**. If we apply an input signal at the inverting terminal (-) then the amplified output signal is 180^o out of phase concerning the applied input signal. If we apply an input signal to the non-inverting terminal (+) then the output signal obtained will be in phase, i.e. it will have no phase shift concerning the input signal.

Power Supply for an Operational Amplifier

As seen from the circuit symbol above it has two input power supply terminals $+V_{CC}$ and $-V_{CC}$. For the operation of an op-amp a dual polarity DC supply is essential. In the dual polarity supply, we connect the $+V_{CC}$ to the positive DC supply and the $-V_{CC}$ terminal to the negative DC supply. However few op-amps can also operate on a single polarity supply. Note that there is no common ground terminal in the op-amps hence the ground has to be established externally.

OPERATIONAL AMPLIFIER STAGES (OP-AMP)

The different stages of a multi-stage operational amplifier are given below.

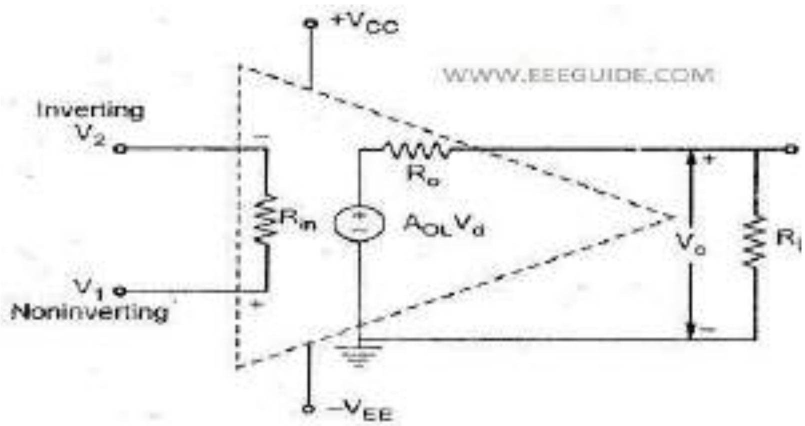


- The op-amp begins with a differential amplifier stage, which operates in the differential mode. Thus the inputs noted with '+' & '- '. The positive sign is for the non-inverting input and negative is for the inverting input.
- The non-inverting input is the ac signal (or dc) applied to the differential amplifier which produces the same polarity of the signal at the output of op-amp. The inverting signals input is the ac signal (or dc) applied to the differential amplifier. This produces 180 degrees out of phase signal at the output.
- The inverting and non-inverting inputs are provided to the input stage which is a dual input, balanced output differential amplifier. The voltage gain required for the amplifier is provided in this stage along with the input resistance for the op-amp.
- The output of the initial stage is given to the intermediate stage, which is driven by the output of the input stage. In this stage direct coupling is used, which makes the dc voltage at the output of the intermediate stage above ground potential.
- Therefore, the dc level at its output must be shifted down to 0Volts with respect to the ground. For this, the level shifting stage is used where usually an emitter follower with the constant current source is applied. The level shifted signal is then given to the output stage where a push-pull amplifier increases the output voltage swing of the signal and also increases the current supplying capability of the op-amp.

EQUIVALENT CIRCUIT OF AN OP-AMP:

The equivalent circuit of an ideal op-amp is shown above. The input voltage V_{DIFF} is the difference voltage (V_1-V_2). Z_{in} is the input impedance and Z_{out} is the output impedance. The gain parameter A is called the open loop gain. If an op-amp does not have any feedback from the output to either of the inputs, it is said to be operating in open-loop configuration.

An ideal op-amp exhibits infinite open loop gain, infinite input impedance, zero output impedance, infinite voltage swing, infinite bandwidth, infinite slew rate and zero input offset voltage.



FIG(A) Equivalent circuit of an op-amp

OPEN LOOP OPERATION OF AN OPERATIONAL AMPLIFIER

op-amp has a differential input and single ended output. So, if we apply two signals one at the inverting and another at the non-inverting terminal, an ideal op-amp will amplify the difference between the two applied input signals. We call this difference between two input signals as the differential input voltage. The equation below gives the output of an operational amplifier.

$$V_{OUT} = A_{OL}(V_1 - V_2)$$

Where, V_{OUT} is the voltage at the output terminal of the op-amp.

A_{OL} is the open-loop gain for the given op-amp and is constant (ideally). For the IC 741 A_{OL} is 2 x 10^5.

V_1 is the voltage at the non-inverting terminal.

V_2 is the voltage at the inverting terminal.

($V_1 - V_2$) is the differential input voltage.

It is clear from the above equation that the output will be non-zero if and only if the differential input voltage is non-zero (V_1 and V_2 are not equal), and will be zero if both V_1 and V_2 are equal.
This is an ideal condition; practically there are small imbalances in the op-amp. The open-loop gain of an op-amp is very high. Hence, an open loop operational amplifier amplifies a small applied differential input voltage to a huge value.
Also, it is true that if we apply small differential input voltage, the operational amplifier amplifies it to a considerable value but this significant value at the output cannot go beyond the supply voltage of the op-amp. Hence it does not violate the law of conservation of energy.

CLOSED LOOP OPERATION (OP-AMP WITH FEEDBACK)

The above-explained operation of the op-amp was for open-loop i.e. without a feedback.
Feedback is introduced in the closed loop configuration. This feedback path feeds the output signal to the input. Hence, at the inputs, two signals are simultaneously present. One of them is the original applied signal, and the other is the feedback signal.

The equation below shows the output of a closed loop op-amp.

$$V_{OUT} = A_{CL}(V_1 - V_2) = A_{CL}V_D$$

Where, V_{OUT} is the voltage at the output terminal of the op-amp.
A_{CL} is the closed loop gain.
The feedback circuit connected to the op-amp determines the closed loop gain A_{CL}.
$V_D = (V_1 - V_2)$ is the differential input voltage.
The feedback as positive if the feedback path feeds the signal from the output terminal back to the non-inverting (+) terminal. Positive feedback is used in oscillators.
The feedback is negative if the feedback path feeds the part of the signal from the output terminal back to the inverting (-) terminal. Negative feedback to the op-amps is used as amplifiers.
Each type of feedback, negative or positive has its advantages and disadvantages.

Positive Feedback ⇒ Oscillator
Negative Feedback ⇒ Amplifier
Th above explanation is the most basic **working principle of operational amplifiers.**

Ideal Op-Amp Characteristics

An ideal op-amp should have the following characteristics:

1. Infinite voltage gain (So that maximum output is obtained)
2. Infinite input resistance (Due to this almost any source can drive it)
3. Zero output resistance (So that there is no change in output due to change in load current)
4. Infinite bandwidth

5. Zero noise
6. Zero power supply rejection ratio (PSSR = 0)
7. Infinite common mode rejection ratio (CMMR = ∞)

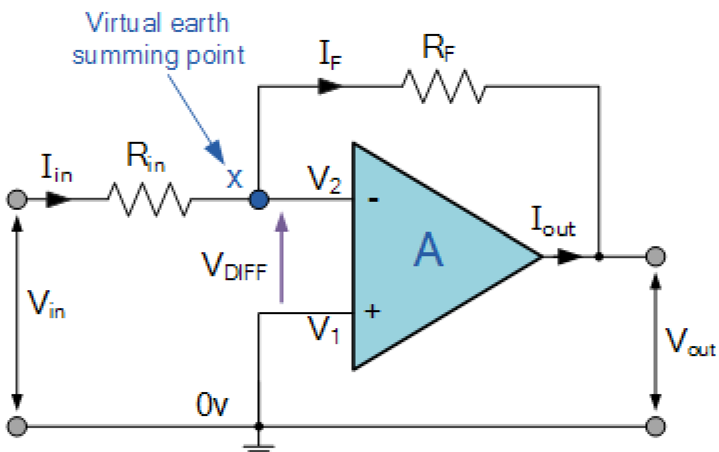
Practical Operational Amplifier

None of the above-given parameters can be practically realized. A practical or real op-amp has some unavoidable imperfections and hence its characteristics differ from the ideal one. A real op-amp will have non-zero and non-infinite parameters.

Applications of Operational Amplifier

The integrated op-amps offer all the advantages of ICs such as high reliability, small size, cheap, less power consumption. They are used in variety of applications such as inverting amplifier and non-inverting amplifiers, unity gain buffer, summing amplifier, differentiator, integrator, adder, instrumentation amplifier, Wien bridge oscillator, Filters etc.

INVERTING OPERATIONAL AMPLIFIER CONFIGURATION



In this Inverting Amplifier circuit the operational amplifier is connected with feedback to produce a closed loop operation. When dealing with operational amplifiers there are two very important rules to remember about inverting amplifiers, these are: "No current flows into the input terminal" and that "V1 always equals V2". However, in real world op-amp circuits both of these rules are slightly broken.

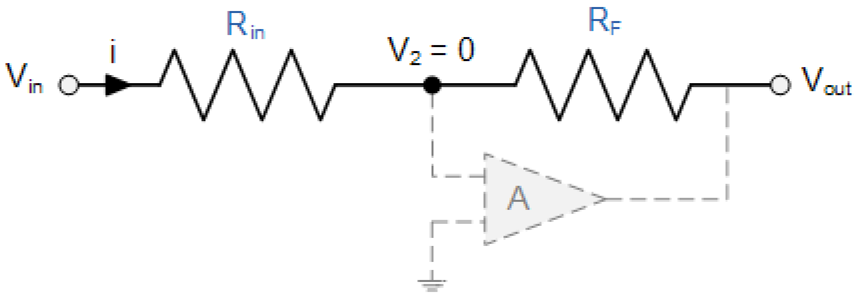
This is because the junction of the input and feedback signal (X) is at the same potential as the positive (+) input which is at zero volts or ground then, the junction is a "Virtual Earth". Because of this virtual earth node the input resistance of the amplifier is equal to the value of the input resistor, Rin and the closed loop gain of the inverting amplifier can be set by the ratio of the two external resistors.

There are two very important rules to remember about Inverting Amplifiers or any operational amplifier for that matter and these are.

- No Current Flows into the Input Terminals
- The Differential Input Voltage is Zero as V1 = V2 = 0 (Virtual Earth)

Then by using these two rules we can derive the equation for calculating the closed-loop gain of an inverting amplifier, using first principles.

Current (i) flows through the resistor network as shown.



$$i = \frac{V_{in}-V_{out}}{R_{in}+R_f}$$

therefore, $i = \frac{Vin-V2}{R_{in}} = \frac{V2-Vout}{R_f}$

$$i = \frac{V_{in}}{R_{in}} - \frac{V_2}{R_{in}} = \frac{V_2}{R_f} - \frac{V_{out}}{R_f}$$

so, $\frac{V_{in}}{R_{in}} = V_2 \left[\frac{1}{R_{in}} + \frac{1}{R_f} \right] - \frac{V_{at}}{R_f}$

and as, $i = \frac{V_{in}-0}{R_{in}} = \frac{0-V_{out}}{R_f} \quad \frac{R_f}{R_{in}} = \frac{0-V_{out}}{V_{in}-0}$

the closed loop Gain (A_v) is given as, $\frac{Vout}{V_{in}} = -\frac{R_f}{R_{in}}$

Then, the **Closed-Loop Voltage Gain** of an Inverting Amplifier is given as.

$$\text{Gain } (\ )_v = \frac{Vout}{V_{in}} = -\frac{R_f}{R_{in}}$$

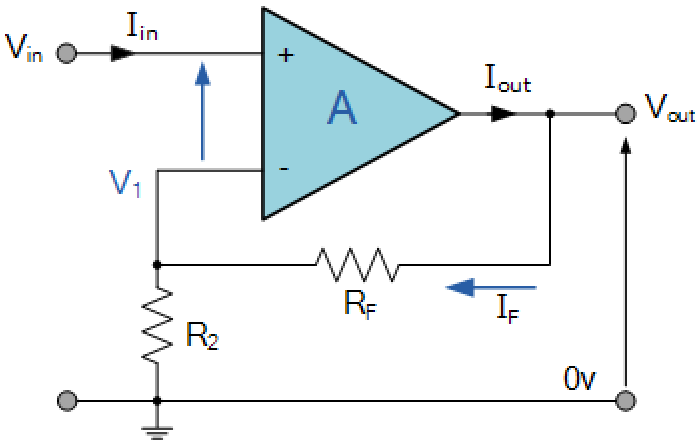
And this can be transposed to give V_{out} as:

$$V_{out} = -\frac{R_f}{R_{in}} \times V_{in}$$

The negative sign in the equation indicates an inversion of the output signal with respect to the input as it is 180° out of phase. This is due to the feedback being negative in value.

The equation for the output voltage V_{out} also shows that the circuit is linear in nature for a fixed amplifier gain as $V_{out} = V_{in}$ x Gain. This property can be very useful for converting a smaller sensor signal to a much larger voltage.

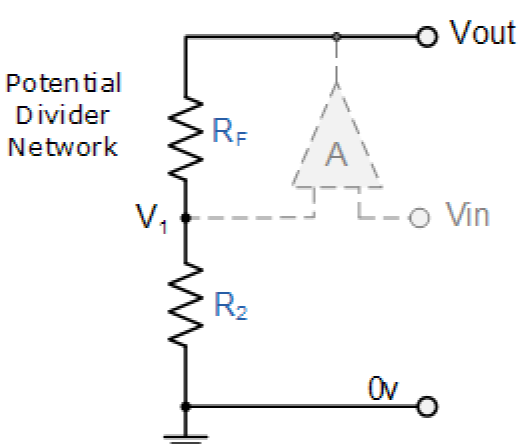
NON-INVERTING OPERATIONAL AMPLIFIER CONFIGURATION



In the previous Inverting Amplifier tutorial, we said that for an ideal op-amp "No current flows into the input terminal" of the amplifier and that "V_1 always equals V_2". This was because the junction of the input and feedback signal ($_1$) are at the same potential.

In other words the junction is a "virtual earth" summing point. Because of this virtual earth node the resistors, R_f and R_2 form a simple potential divider network across the non-inverting amplifier with the voltage gain of the circuit being determined by the ratios of $R_2$and$R_f$as shown below.

Equivalent Potential Divider Network



Then using the formula to calculate the output voltage of a potential divider network, we can calculate the closed-loop voltage gain (A_V) of the **Non-inverting Amplifier** as follows:

$$V_1 = \frac{2}{R_2 + R_f} \text{x}\, V_{out}$$

Ideal summing point: $V_1 = V_{in}$

Voltage Gain, A_V is equal to : $\frac{Vout}{V_{in}}$

Then, $A_v = \frac{V_{out}}{V_{in}} = \frac{R_2 + R_f}{R_2}$

Transpose to give: $A_V = \frac{Vout}{V_{in}} = 1 + \frac{R_f}{R_2}$

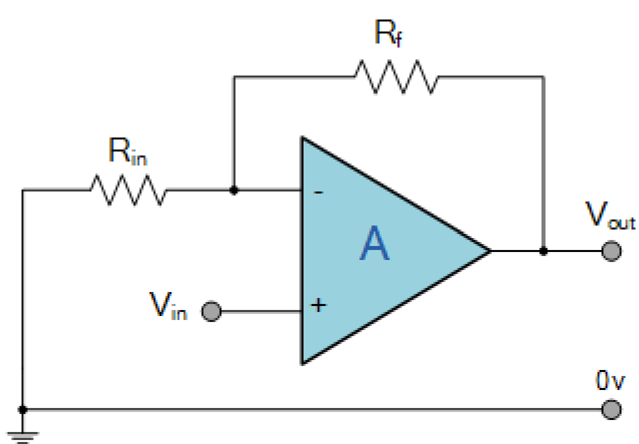
Then the closed loop voltage gain of a **Non-inverting Operational Amplifier** will be given as:

$$A_V = \frac{Vout}{V_{in}} = 1 + \frac{R_f}{R_2}$$

We can see from the equation above that the overall closed-loop gain of a non-inverting amplifier will always be greater but never less than one (unity), it is positive in nature and is determined by the ratio of the values of Rƒ and R2.

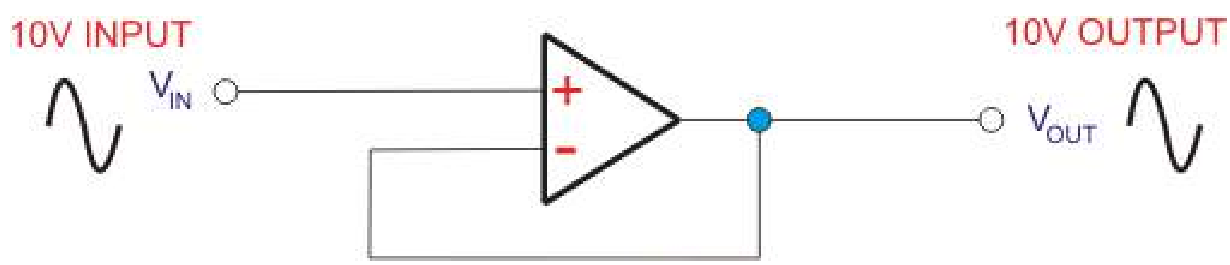
If the value of the feedback resistor Rƒ is zero, the gain of the amplifier will be exactly equal to one (unity). If resistor R2 is zero the gain will approach infinity, but in practice it will be limited to the operational amplifiers open-loop differential gain, (A_O).

We can easily convert an inverting operational amplifier configuration into a non-inverting amplifier configuration by simply changing the input connections as shown.



VOLTAGE FOLLOWER AND BUFFER

Voltage follower is an Op-amp circuit whose output voltage straight away follows the input voltage. That is output voltage is equivalent to the input voltage. Op-amp circuit does not provide any amplification. Thus, voltage gain is equal to 1. They are similar to discrete emitter follower. The other names of voltage follower are Isolation Amplifier, Buffer Amplifier, and Unity-Gain Amplifier. The voltage follower provides no attenuation or no amplification but only buffering. This circuit has an advantageous characteristic of very high input impedance. This high input impedance of voltage follower is the reason of it being used in several circuits. The **voltage follower** gives an efficient isolation of output from the input signal. The circuit of voltage follower is shown below.



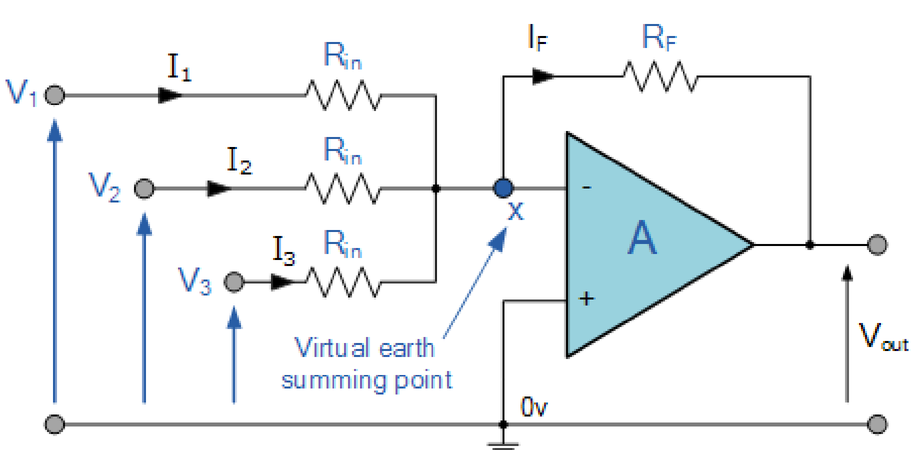
We all know the most fundamental law; that is Ohm's law.

$$Current = \frac{Voltage}{Resistance}$$

So, we can say that when resistance increases, the current drawn from the power source decreases. Thus, we conclude that the power is unaffected if the current is feeding a load of high impedance.

DIFFERENTIAL AMPLIFIER 8.9.1

Summing Amplifier Circuit



In this simple summing amplifier circuit, the output voltage, (Vout) now becomes proportional to the sum of the input voltages, V_1, V_2, V_3, etc. Then we can modify the original equation for the inverting amplifier to take account of these new inputs thus:

$$I_F = I_1 + I_2 + I_3 = -\left[\frac{1}{R_{in}} + \frac{V2}{R_{in}} + \frac{V3}{R_{in}}\right]$$

Inverting equation: $V_{out} = -\frac{R_f}{R_{in}} \times V_{in}$

Then, $-V_{out} = \left[\frac{R_f}{R_{in}} V_1 + \frac{R_f}{R_{in}} V_2 + \frac{R_f}{R_{in}} V_3\right]$

However, if all the input impedances, (R_{IN}) are equal in value, we can simplify the above equation to give an output voltage of:

Summing Amplifier Equation

$$-V_{out} = \frac{R_f}{R_{in}} (V_1 + V_2 + V_3)$$

An operational amplifier circuit will amplify each individual input voltage and produce an output voltage signal that is proportional to the algebraic “SUM” of the three individual input voltages V_1, V_2 and V_3.

This is because the input signals are effectively isolated from each other by the “virtual earth” node at the inverting input of the op-amp. A direct voltage addition can also be obtained when all the resistances are of equal value and Rf is equal to Rin.

When the summing point is connected to the inverting input of the op-amp the circuit will produce the negative sum of any number of input voltages. Likewise, when the summing point is connected to the non-inverting input of the op-amp, it will produce the positive sum of the input voltages.

A **Scaling Summing Amplifier** can be made if the individual input resistors are “NOT” equal. Then the equation would have to be modified to:

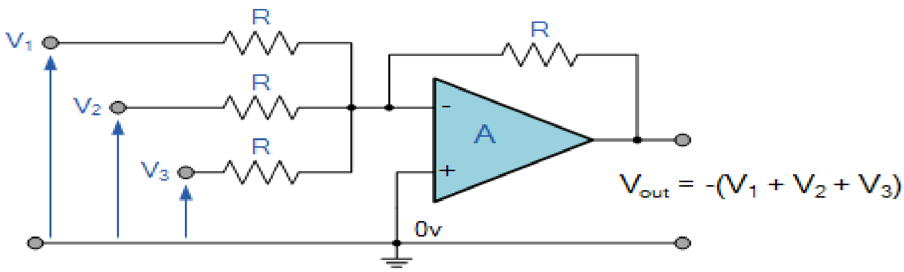
$$-V_{out} = V_1 \left(\frac{R_f}{R_1}\right) + V_2 \left(\frac{R_f}{R_2}\right) + V_3 \left(\frac{R_f}{R_3}\right)$$

We can rearrange the above formula to make the feedback resistor Rf the subject of the equation giving the output voltage as:

$$-V_{out} = R_f \left(\frac{1}{R_1} + \frac{V2}{R_2} + \frac{V3}{R_3}\right)$$

This allows the output voltage to be easily calculated if more input resistors are connected to the amplifiers inverting input terminal. The input impedance of each individual channel is the value of their respective input resistors, ie, R_1, R_2, R_3 ... etc.

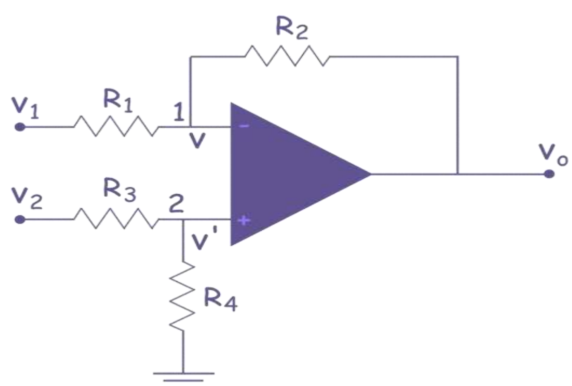
Sometimes we need a summing circuit to just add together two or more voltage signals without any amplification. By putting all of the resistances of the circuit above to the same value R, the op-amp will have a voltage gain of unity and an output voltage equal to the direct sum of all the input voltages as shown:



The **Summing Amplifier** is a very flexible circuit indeed, enabling us to effectively "Add" or "Sum" (hence its name) together several individual input signals. If the inputs resistors, R_1, R_2, R_3 etc, are all equal a "unity gain inverting adder" will be made. However, if the input resistors are of different values a "scaling summing amplifier" is produced which will output a weighted sum of the input signals.

Op-amp Sub tractor Circuit:

A difference amplifier or op amp subtractor is a specially designed op amp based amplifier circuit, which amplifies the difference between two input signals and rejects any signals common to both inputs.



Let us consider the above op amp circuit. Now, by applying Kirchhoff Current Law at node 1, we get,

$$\frac{v_1 - v}{R_1} = \frac{v - v_0}{R_2}$$

We have written this equation by assuming that there is no current entering in the inverting terminal of the op amp.

Now, by simplifying the above equation, we get,

$$\frac{v_1}{R_1} + \frac{v_0}{R_2} = \frac{v}{R_1} + \frac{v}{R_2}$$

$$\Rightarrow \frac{v_0}{R_2} = \frac{v}{R_1} + \frac{v}{R_2} - \frac{v_1}{R_1}$$

$$\Rightarrow v_0 = \left(\frac{R_2}{R_1} + 1\right) v - \frac{R_2}{R_1} v_1 \Rightarrow v_0 = \left(1 + \frac{R_2}{R_1}\right) v - \frac{R_2}{R_1} v_1 \quad \ldots\ldots (i)$$

Now, by applying Kirchhoff Current Law, at node 2, we get,

$$\frac{v_2 - v'}{R_3} = \frac{v' - 0}{R_4} \;\Rightarrow\; v_2 = v'\left(1 + \frac{R_3}{R_4}\right)$$

$$\Rightarrow \frac{v_2}{R_3} = \frac{v'}{R_4} + \frac{v'}{R_3} \Rightarrow\; v' = v_2\left(\frac{R_4}{R_3 + R_4}\right) \;\; \ldots\ldots(ii)$$

We know that, in ideal op amp, voltage at inverting input is same as the voltage at non inverting input. Hence,

$$v = v'$$

So, now from equation (i) and (ii), we get,

$$v_0 = \left(1 + \frac{R_2}{R_1}\right)\frac{R_4}{R_3 + R_4}v_2 - \frac{R_2}{R_1}v_1$$

$$\Rightarrow\; v_0 = \frac{\left(1 + \frac{R_2}{R_1}\right)}{\left(1 + \frac{R_3}{R_4}\right)}v_2 - \frac{R_2}{R_1}v_1$$

$$v_0 = \frac{R_2}{R_1} \cdot \frac{1 + \frac{R_1}{R_2}}{1 + \frac{R_3}{R_4}}v_2 - \frac{R_2}{R_1}v_1 \cdots\cdots\cdots(iii)$$

The **difference amplifier** must reject any signal common to both inputs. That means, if polarity and magnitude of both input signals are same, the output must be zero.

$$\because\; when,\; v_1 = v_2,\; then,\; v_0 = 0$$

This condition must be satisfied only when,

$$\frac{R_1}{R_2} = \frac{R_3}{R_4} \Leftrightarrow R_1 : R_2 = R_3 : R_4$$

In that case, equation (iii) becomes,

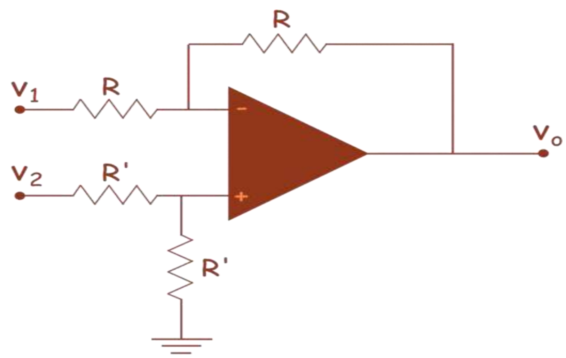
$$v_0 = \frac{R_2}{R_1}v_2 - \frac{R_2}{R_1}v_1$$

$$\Rightarrow\; v_0 = \frac{R_2}{R_1}(v_2 - v_1) \;\; \ldots\ldots(iv)$$

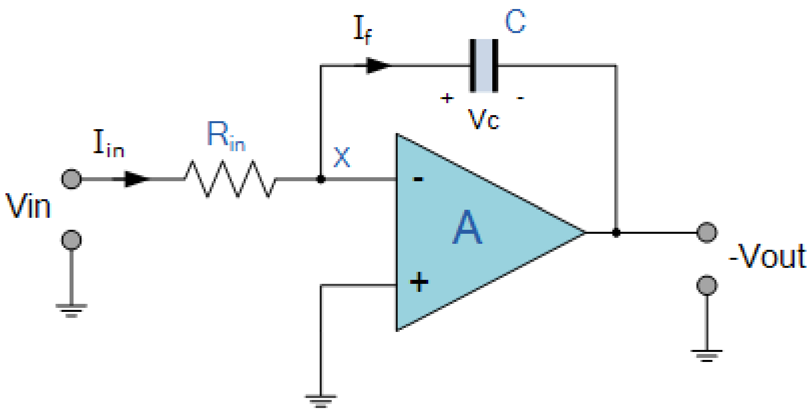
Again, if we make, $R_1 = R_2$, then equation (iv) becomes,

$$v_0 = v_2 - v_1$$

So, if $R_1 = R_2$ and also $R_3 = R_4$ then the difference amplifier becomes a perfect subtractor, which subtracts directly the input signals.
Finally, the circuit of **op amp substractor** becomes,



Op-amp Integrator Circuit



As its name implies, the **Op-amp Integrator** is an operational amplifier circuit that performs the mathematical operation of **integration** that is we can cause the output to respond to changes in the input voltage over time as the op-amp integrator produces an *output voltage which is proportional to the integral of the input voltage*.

When a step voltage, Vin is firstly applied to the input of an integrating amplifier, the uncharged capacitor C has very little resistance and acts a bit like a short circuit allowing maximum current to flow via the input resistor, Rin as potential difference exists between the two plates. No current flows into the amplifiers input and point X is a virtual earth resulting in zero output. As the impedance of the capacitor at this point is very low, the gain ratio of X_C/R_{IN} is also very small giving an overall voltage gain of less than one, (voltage follower circuit).

As the feedback capacitor, C begins to charge up due to the influence of the input voltage, its impedance Xc slowly increase in proportion to its rate of charge. The capacitor charges up at a rate determined by the RC time constant, (τ) of the series RC network. Negative feedback forces the op-amp to produce an output voltage that maintains a virtual earth at the op-amp's inverting input.

Since the capacitor is connected between the op-amp's inverting input (which is at virtual ground potential) and the op-amp's output (which is now negative), the potential voltage, Vc developed across the capacitor slowly increases causing the charging current to decrease as the impedance of the capacitor increases. This results in the ratio of Xc/Rin increasing producing a linearly increasing ramp output voltage that continues to increase until the capacitor is fully charged.

At this point the capacitor acts as an open circuit, blocking any more flow of DC current. The ratio of feedback capacitor to input resistor (X_C/R_{IN}) is now infinite resulting in infinite gain. The result of this high gain (similar to the op-amps open-loop gain), is that the output of the amplifier goes into saturation as shown below. (Saturation occurs when the output voltage of the amplifier swings heavily to one voltage supply rail or the other with little or no control in between).

The rate at which the output voltage increases (the rate of change) is determined by the value of the resistor and the capacitor, "RC time constant".

We know that the voltage on the plates of a capacitor is equal to the charge on the capacitor divided by its capacitance giving Q/C. Then the voltage across the capacitor is output Vout therefore: -Vout = Q/C. If the capacitor is charging and discharging, the rate of charge of voltage across the capacitor is given as:

$$V_c = \frac{\ }{C},\ V_c = V_x - V_{out} = 0 - V_{out}$$

$$-\frac{dV_{out}}{dt} = \frac{d}{Cdt} = \frac{1}{C}\frac{dQ}{dt}$$

But dQ/dt is electric current and since the node voltage of the integrating op-amp at its inverting input terminal is zero, X = 0, the input current I(in) flowing through the input resistor, Rin is given as:

$$I_{in} = \frac{V_{in}-0}{R_{in}} = \frac{V_{in}}{R_{in}}$$

The current flowing through the feedback capacitor C is given as:

$$I_f = C\frac{dV_{out}}{dt} = C\frac{dQ}{Cdt} = \frac{dQ}{dt} = \frac{dV_{ou}\,.\,C}{dt}$$

Assuming that the input impedance of the op-amp is infinite (ideal op-amp), no current flows into the op-amp terminal. Therefore, the nodal equation at the inverting input terminal is given as:

$$I_{in} = I_f = \frac{V_{in}}{R_{in}} = \frac{dV_{out}\,.C}{dt}$$

$$\frac{V_{in}}{V_{out}} \text{ x } \frac{dt}{R_{in}\,C} = 1$$

From which we derive an ideal voltage output for the **Op-amp Integrator** as:

$$V_{out} = -\frac{1}{R_{in}\,C}\int_0^t V_{in}\ dt = -\int_0^t V_{in}\ \frac{dt}{R_{in}\,.C}$$

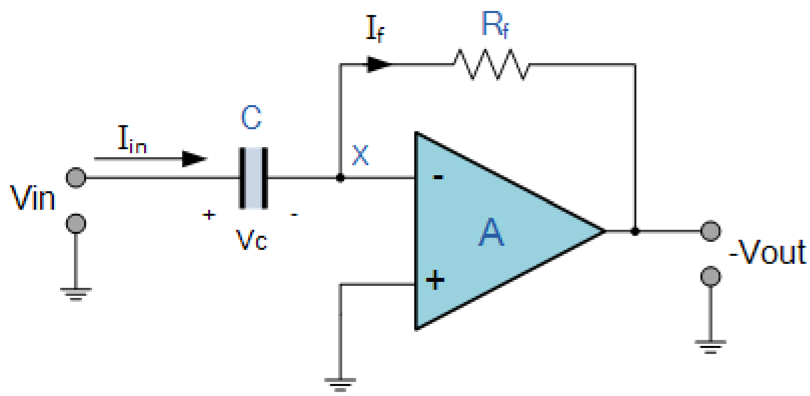
To simplify the math's a little, this can also be re-written as:

$$V_{out} = -\frac{1}{j\omega RC} V_{in}$$

Where: ω = 2πf and the output voltage Vout is a constant 1/RC times the integral of the input voltage V_{IN} with respect to time.

Thus the circuit has the transfer function of an inverting integrator with the gain constant of -1/RC. The minus sign (–) indicates a 180° phase shift because the input signal is connected directly to the inverting input terminal of the operational amplifier.

Op-amp Differentiator Circuit



The input signal to the differentiator is applied to the capacitor. The capacitor blocks any DC content so there is no current flow to the amplifier summing point, X resulting in zero output voltage. The capacitor only allows AC type input voltage changes to pass through and whose frequency is dependant on the rate of change of the input signal.

At low frequencies the reactance of the capacitor is "High" resulting in a low gain (Rƒ/Xc) and low output voltage from the op-amp. At higher frequencies the reactance of the capacitor is much lower resulting in a higher gain and higher output voltage from the differentiator amplifier.

However, at high frequencies an op-amp differentiator circuit becomes unstable and will start to oscillate. This is due mainly to the first-order effect, which determines the frequency response of the op-amp circuit causing a second-order response which, at high frequencies gives an output voltage far higher than what would be expected. To avoid this the high frequency gain of the circuit needs to be reduced by adding an additional small value capacitor across the feedback resistor Rƒ.

Since the node voltage of the operational amplifier at its inverting input terminal is zero, the current, i flowing through the capacitor will be given as:

$$I_{in} = I_f \text{ and } I_f = -\frac{Vout}{R_f}$$

The charge on the capacitor equals Capacitance times Voltage across the capacitor

$$Q = C \times V_{in}$$

Thus the rate of change of this charge is:

$$\frac{dQ}{dt} = C\frac{dV_{in}}{dt}$$

But dQ/dt is the capacitor current, *i*

$$I_{in} = C\frac{dV_{in}}{dt} = I_f$$

$$-\frac{V_{out}}{R_f} = C\frac{dV_{in}}{dt}$$

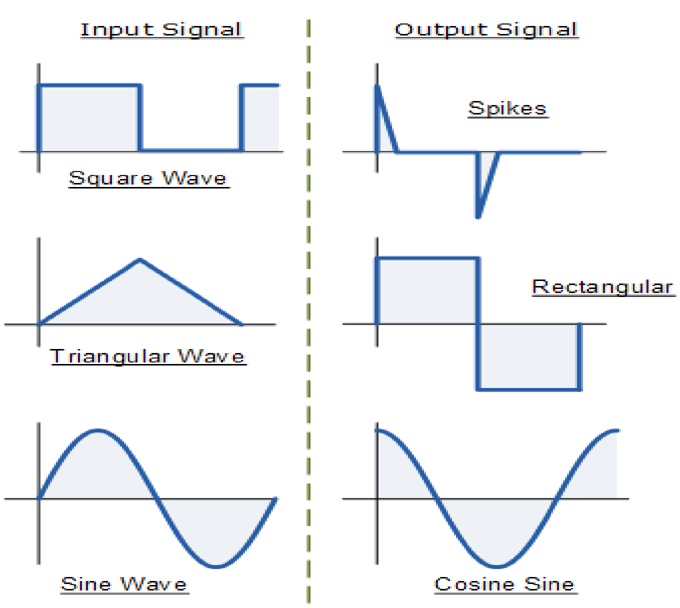
From which we have an ideal voltage output for the op-amp differentiator is given as:

$$V_{out} = -R_f\, C\frac{dV_{in}}{dt}$$

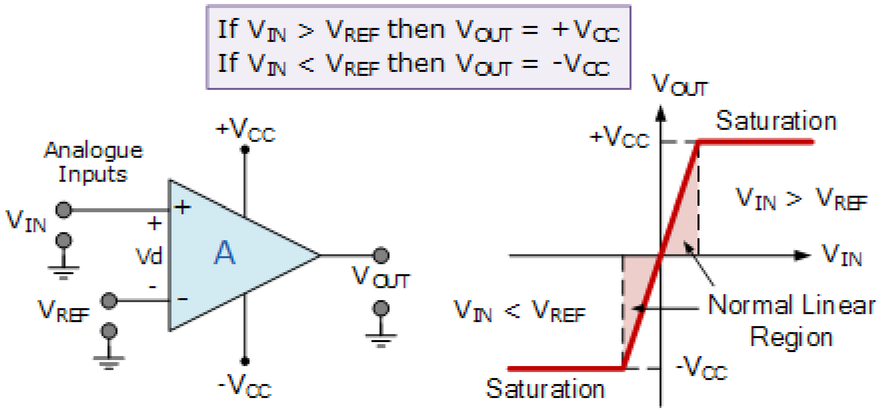
Therefore, the output voltage V_{out} is a constant $-R_f$x C times the derivative of the input voltage Vin with respect to time. The minus sign (–) indicates a 180° phase shift because the input signal is connected to the inverting input terminal of the operational amplifier.

One final point to mention, the **Op-amp Differentiator** circuit in its basic form has two main disadvantages compared to the previous operational amplifier integrator circuit. One is that it suffers from instability at high frequencies as mentioned above, and the other is that the capacitive input makes it very susceptible to random noise signals and any noise or harmonics present in the source circuit will be amplified more than the input signal itself. This is because the output is proportional to the slope of the input voltage so some means of limiting the bandwidth in order to achieve closed-loop stability is required.

Op-amp Differentiator Waveforms



Op-amp Comparator Circuit



With reference to the op-amp comparator circuit above, lets first assume that V_{IN} is less than the DC voltage level at V_{REF}, ($V_{IN} < V_{REF}$). As the non-inverting (positive) input of the comparator is less than the inverting (negative) input, the output will be LOW and at the negative supply voltage, -Vcc resulting in a negative saturation of the output.

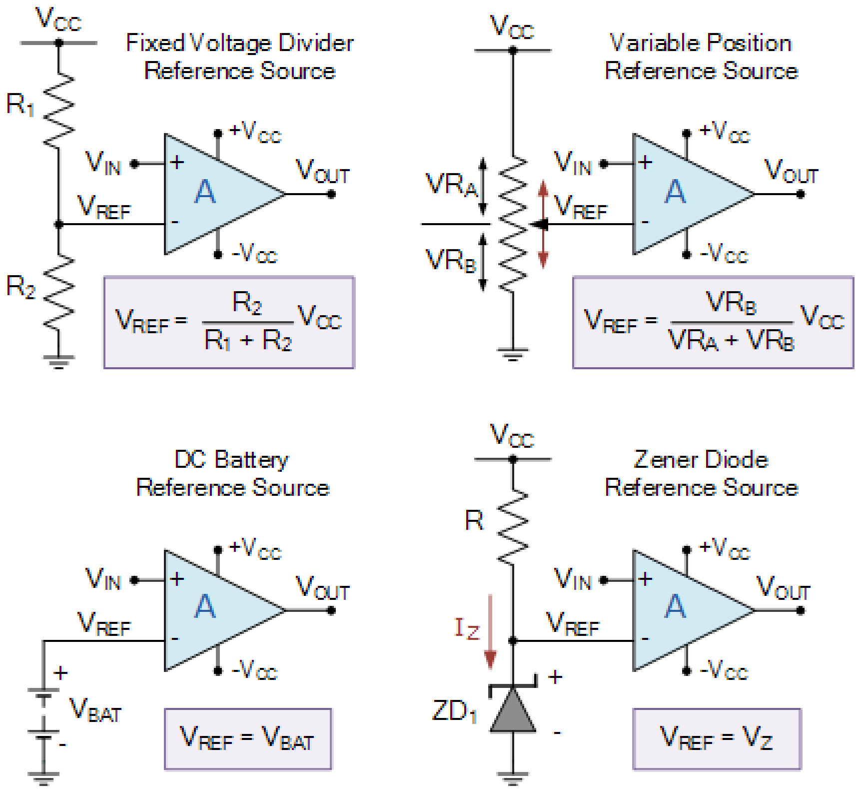
If we now increase the input voltage, V_{IN} so that its value is greater than the reference voltage V_{REF} on the inverting input, the output voltage rapidly switches HIGH towards the positive supply voltage, +Vcc resulting in a positive saturation of the output. If we reduce again the input voltage V_{IN}, so that it is slightly less than the reference voltage, the op-amp's output switches back to its negative saturation voltage acting as a threshold detector.

Op-amp voltage comparator is a device whose output is dependent on the value of the input voltage, V_{IN} with respect to some DC voltage level as the output is HIGH when the voltage on the non-inverting input is greater than the voltage on the inverting input, and LOW when the non-inverting input is less than the inverting input voltage. This condition is true regardless of whether the input signal is connected to the inverting or the non-inverting input of the comparator.

The value of the output voltage is completely dependent on the op-amps power supply voltage.

$$V_{OUT} = +V_{cc} \text{ or } V_{OUT} = -V_{cc}.$$

The basic op-amp comparator produces a positive or negative voltage output by comparing its input voltage against some preset DC reference voltage. Generally, a resistive voltage divider is used to set the input reference voltage of a comparator, but a battery source, Zener diode or potentiometer for a variable reference voltage can all be used as shown.



Comparator Reference Voltages

In theory the comparators reference voltage can be set to be anywhere between 0v and the supply voltage but there are practical limitations on the actual voltage range depending on the op-amp comparator being device used.

GANDHI INSTITUTE OF EXCELLENT TECHNOCRATS, BHUBANESWAR

DEPARTMENT OF ELECTRICAL ENGINEERING

LECTURE NOTES

Year & Semester: 2nd Year, 4th Semester (EE)

Subject Code/Name: *TH-4, Generation Transmission & Distribution (part-I)*

Prepared By:-

COURSE COTENTS

1. GENERATION OF ELECTRICITY

Elementary idea on generation of electricity from Thermal,, Hydro,Nuclear,Power station.

Introduction to Solar Power Plant (Photovoltaic cells).

Layout diagram of generating stations.

2. TRANSMISSION OF ELECTRIC POWER

Layout of transmission and distribution scheme.

Voltage Regulation & efficiency of transmission.

State and explain Kelvin's law for economical size of conductor.

Corona and corona loss on transmission lines.

3. OVER HEAD LINES

Types of supports, size and spacing of conductor.

Types of conductor materials.

State types of insulator and cross arms.

Sag in overhead line with support at same level and different level.
(approximate formula effect of wind, ice and temperature on sag)

Simple problem on sag.

4. PERFORMANCE OF SHORT & MEDIUM LINES

4.1. Calculation of regulation and efficiency.

5. EHV TRANSMISSION

EHV AC transmission.

5.1..1. Reasons for adoption of EHV AC transmission.

5.1..2. Problems involved in EHV transmission.

HV DC transmission.

5.2..1. Advantages and Limitations of HVDC transmission system.

6. DISTRIBUTION SYSTEMS

Introduction to Distribution System.

Connection Schemes of Distribution System: (Radial, Ring Main and Inter connected system)

DC distributions.

Distributor fed at one End.

Distributor fed at both the ends.

Ring distributors.

AC distribution system.

Method of solving AC distribution problem.

Three phase four wire star connected system arrangement.

7. UNDERGROUND CABLES

Cable insulation and classification of cables.

Types of L. T. & H.T. cables with constructional features.

Methods of cable lying.

Localization of cable faults: Murray and Varley loop test for short circuit fault / Earth fault.

8. ECONOMIC ASPECTS

Causes of low power factor and methods of improvement of power factor in power system.

Factors affecting the economics of generation: (Define and explain)

- Load curves.
- Demand factor.
- Maximum demand.
- Load factor.
- Diversity factor.
- Plant capacity factor.

Peak load and Base load on power station.

9. TYPES OF TARIFF

Desirable characteristic of a tariff.

Explain flat rate, block rate, two part and maximum demand tariff. (Solve Problems)

10. SUBSTATION

Layout of LT, HT and EHT substation.

Earthing of Substation, transmission and distribution lines

COURSE OUTCOMES:-

1. Analyze the performance of various units involved in the power plant and apply power system fundamental to design of a system the meets specific needs. Acquire the knowledge of transmission and distribution scheme.

2. Design a power system solution based on the problem requirements and realistic constraints by calculation of sag, corona etc.

3. Calculation of voltage regulation and transmission efficiency by the performance of short and medium transmission lines.

4. Acquire the knowledge of EHVAC and HVDC transmissions systems method of underground cable lying and to locate the cable fault.

5. Identify the different types of DC and AC distribution system and calculate various parameters. Understand the factors affecting the economics of generation and acquire the knowledge of traffic .

6. Explain lay out of LT,HT,EHT substations and can be able to understand earthing of these system .

Lecture note

Semester- 4th

Subject – GENERATION TRANSMISSION AND DISTRIBUTION

Subject code- Th4

Branch- Electrical Engineering

Chapter1 Generation of electricity

IMPORTANCE OF ELECTRICAL ENERGY:

Energy may be needed as heat, as light, as motive power etc. The present-day advancement in science and technology has made it possible to convert electrical energy into any desired form. This has given electrical energy a place of pride in the modern world. The survival of industrial undertakings and our social structures depends primarily upon low cost and uninterrupted supply of electrical energy. In fact, the advancement of a country is measured in terms of per capita consumption of electrical energy. Electrical energy is superior to all other forms of energy due to the following reasons:

(i) Convenient form. Electrical energy is a very convenient form of energy. It can be easily converted into other forms of energy. For example, if we want to convert electrical energy into heat, the only thing to be done is to pass electrical current through a wire of high resistance e.g., a heater. Similarly, electrical energy can be converted into light (e.g. electric bulb), mechanical energy (e.g. electric motors) etc.

(ii) Easy control. The electrically operated machines have simple and convenient starting, control and operation. For instance, an electric motor can be started or stopped by turning on or off a switch. Similarly, with simple arrangements, the speed of electric motors can be easily varied over the desired range.

(iii) Greater flexibility. One important reason for preferring electrical energy is the flexibility that it offers. It can be easily transported from one place to another with the help of conductors.

(iv) Cheapness. Electrical energy is much cheaper than other forms of energy. Thus it is overall economical to use this form of energy for domestic, commercial and industrial purposes.

(v) Cleanliness. Electrical energy is not associated with smoke, fumes or poisonous gases. Therefore, its use ensures cleanliness and healthy conditions.

(vi) High transmission efficiency. The consumers of electrical energy are generally situated quite away from the centres of its production. The electrical energy can be transmitted conveniently and efficiently from the centres of generation to the consumers with the help of overhead conductors known as transmission lines.

The conversion of energy available in different forms in nature into electrical energy is known as generation of electrical energy. Electrical energy is a manufactured commodity like clothing, furniture or tools. Just as the manufacture of a commodity involves the conversion of raw materials available in nature into the desired form, similarly electrical energy is produced from the forms of energy available in nature. However, electrical energy differs in one important respect. Whereas other commodities may be produced at will and consumed as needed, the electrical energy must be produced and transmitted to the point of use at the instant it is needed. The entire process takes only a fraction of a second. This instantaneous production of electrical energy introduces technical and economic considerations unique to the electrical power industry. Energy is available in various forms from different natural sources such as pressure head of water, chemical energy of fuels, nuclear energy of radioactive substances etc. All these forms of energy can be converted into electrical energy by the use of suitable arrangements. The arrangement essentially employs (see Fig. 1.1) an alternator coupled to a prime mover. The prime mover is driven by the energy obtained from various sources such as burning of fuel, pressure of water, force of wind etc. For example, chemical energy of a fuel (e.g., coal) can be used to produce steam at high temperature and pressure. The steam is fed to a prime mover which may be a steam engine or a steam turbine. The turbine converts heat energy of steam into mechanical energy which is further converted into electrical energy by the alternator. Similarly, other forms of energy can be converted into electrical energy by employing suitable machinery and equipment.

Generating Stations Bulk electric power is produced by special plants known as generating stations or power plants. A generating station essentially employs a primemover coupled to an alternator for the production of electric power. The prime mover (e.g., steam turbine, water turbine etc.) converts energy from some other form into mechanical energy. The alternator converts mechanical energy of the prime mover into electrical energy. The electrical energy produced by the generating station is transmitted and distributed with the help of conductors to various consumers. It may be emphasized here that apart from prime mover-alternator combination, a modern generating station employs several auxiliary equipment and instruments to ensure cheap, reliable and continuous service. Depending upon the form of energy converted into electrical energy, the generating stations are classified as under:

(i) Steam power stations
(ii) Hydroelectric power stations
(iii) Diesel power stations
(iv) Nuclear power stations

Steam power station:

Steam Power Station (Thermal Station) A generating station which converts heat energy of coal combustion into electrical energy is known as a steam power station. A steam power station basically works on the Rankine cycle. Steam is produced in the boiler by utilizing the heat of coal combustion. The steam is then expanded in the prime mover (i.e., steam turbine) and is condensed in a condenser to be fed into the boiler again. The steam turbine drives the alternator which converts mechanical energy of the turbine into electrical energy. This type of power station is suitable where coal and water are available in abundance and a large amount of electric power is to be generated.

Advantages

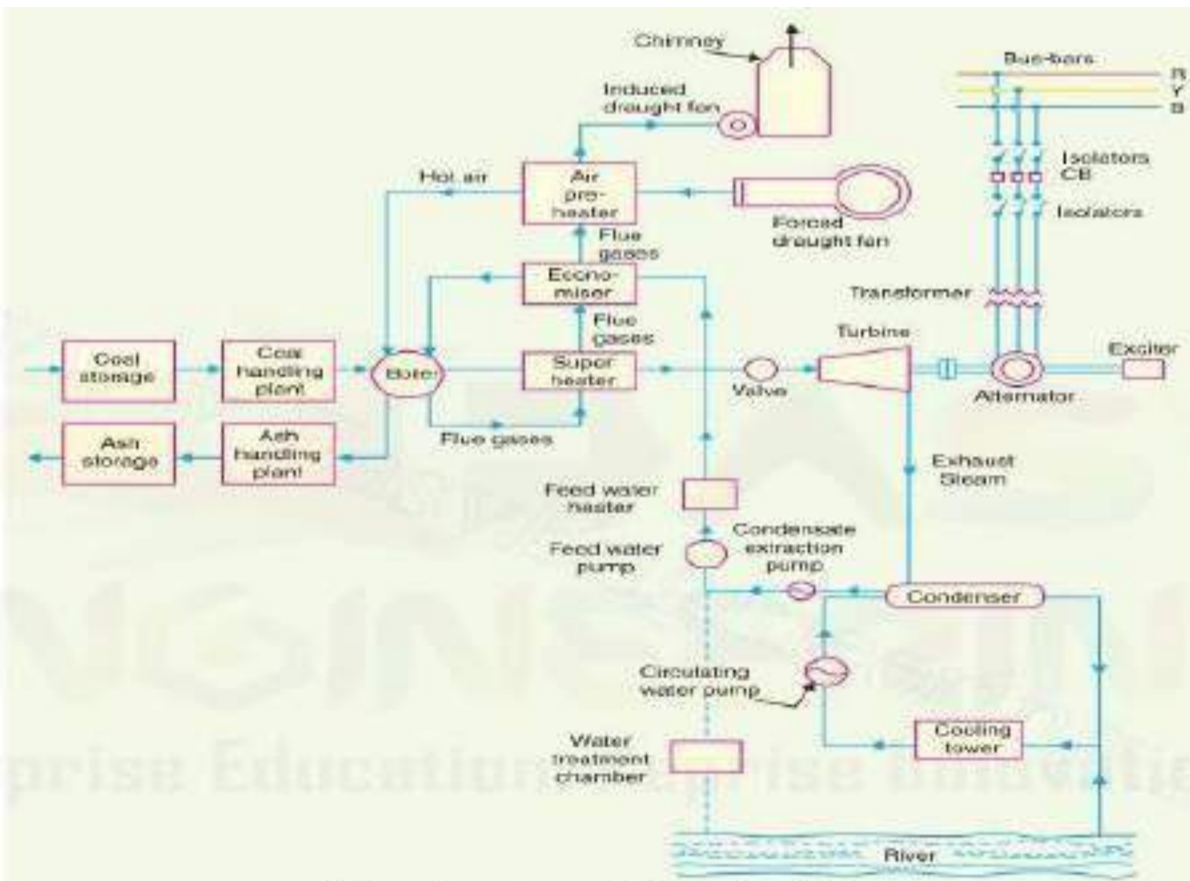
(i) The fuel (i.e., coal) used is quite cheap.
(ii) Less initial cost as compared to other generating stations.
(iii) It can be installed at any place irrespective of the existence of coal. The coal can be transported to the site of the plant by rail or road.
(iv) It requires less space as compared to the hydroelectric power station.
(v) The cost of generation is lesser than that of the diesel power station.

Disadvantages

(i) It pollutes the atmosphere due to the production of large amount of smoke and fumes.
(ii) It is costlier in running cost as compared to hydroelectric plant.

Schematic Arrangement of Steam Power Station:

Although steam power station simply involves the conversion of heat of coal combustion into electrical energy, yet it embraces many arrangements for proper working and efficiency. The schematic arrangement of a modern steam power station is shown in Fig. 2.1.



Schematic arrangement of Steam Power Station
Fig. 2.1

The whole arrangement can be divided into the following stages for the sake of simplicity :

1. Coal and ash handling arrangement

2. Steam generating plant

3. Steam turbine

4. Alternator

5. Feed water

6. Cooling arrangement

Coal and ash handling plant:

The coal is transported to the power station by road or rail and is stored in the coal storage plant. Storage of coal is primarily a matter of protection against coal strikes, failure of transportation system and general coal shortages. From the coal storage plant, coal is delivered to the coal handling plant where it is pulverised (i.e., crushed into small pieces) in order to increase its surface exposure, thus promoting rapid combustion without using large quantity of excess air. The pulverised coal is fed to the boiler by belt conveyors. The coal is burnt in the boiler and the ash produced after the complete combustion of coal is removed to the ash handling plant and then delivered to the ash storage plant for disposal. The removal of the ash from the boiler furnace is necessary for proper burning of coal. It is worthwhile to give a passing reference to the amount of coal burnt and ash produced in a modern thermal power station. A 100 MW station operating at 50% load factor may burn about 20,000 tons of coal per month and ash produced may be to the tune of 10% to 15% of coal fired i.e., 2,000 to 3,000 tons. In fact, in a thermal station, about 50% to 60% of the total operating cost consists of fuel purchasing and its handling.

Steam generating plant:

The steam generating plant consists of a boiler for the production of steam and other auxiliary equipment for the utilisation of flue gases.

(i) **Boiler.** The heat of combustion of coal in the boiler is utilised to convert water into steam at high temperature and pressure. The flue gases from the boiler make their journey through superheater, economiser, air pre-heater and are finally exhausted to atmosphere through the chimney.

(ii) **Superheater.** The steam produced in the boiler is wet and is passed through a superheater where it is dried and superheated (i.e., steam temperature increased above that of boiling point of water) by the flue gases on their way to chimney. Superheating provides two principal benefits. Firstly, the overall efficiency is increased. Secondly, too much condensation in the last stages of turbine (which would cause blade corrosion) is avoided. The superheated steam from the superheater is fed to steam turbine through the main valve.

(iii) **Economiser.** An economiser is essentially a feed water heater and derives heat from the flue gases for this purpose. The feed water is fed to the economiser before supplying to the boiler. The economiser extracts a part of heat of flue gases to increase the feed water temperature.

(iv) **Air preheater.** An air preheater increases the temperature of the air supplied for coal burning by deriving heat from flue gases. Air is drawn from the atmosphere by a forced draught fan and is passed through air preheater before supplying to the boiler furnace. The air preheater extracts heat from flue gases and increases the temperature of air used for coal combustion. The principal benefits of preheating the air are: increased thermal efficiency and increased steam capacity per square metre of boiler surface.

Steam turbine:

The dry and superheated steam from the superheater is fed to the steam turbine through main valve. The heat energy of steam when passing over the blades of turbine is converted into mechanical energy. After giving heat energy to the turbine, the steam is exhausted to the condenser which condenses the exhausted steam by means of cold-water circulation.

Alternator:

The steam turbine is coupled to an alternator. The alternator converts mechanical energy of turbine into electrical energy. The electrical output from the alternator is delivered to the bus bars through transformer, circuit breakers and isolators.

Feed water:

The condensate from the condenser is used as feed water to the boiler. Some water may be lost in the cycle which is suitably made up from external source. The feed water on its way to the boiler is heated by water heaters and economiser. This helps in raising the overall efficiency of the plant.

Cooling arrangement:

In order to improve the efficiency of the plant, the steam exhausted from the turbine is condensed* by means of a condenser. Water is drawn from a natural source of supply such as a river, canal or lake and is circulated through the condenser. The circulating water takes up the heat of the exhausted steam and itself becomes hot. This hot water coming out from the condenser is discharged at a suitable location down the river. In case the availability of water from the source of supply is not assured throughout the year, cooling towers are used. During the scarcity of water in the river, hot water from the condenser is passed on to the cooling towers where it is cooled. The cold water from the cooling tower is reused in the condenser.

Equipment of Steam Power Station A modern steam power station is highly complex and has numerous equipment and auxiliaries. However, the most important constituents of a steam power station are:

1. Steam generating equipment

2. Condenser

3. Prime mover

4. Water treatment plant

5. Electrical equipment.

Hydro power plant:

A generating station which utilises the potential energy of water at a high level for the generation of electrical energy is known as a **hydro-electric power station.**

Hydro-electric power stations are generally located in hilly areas where dams can be built conveniently and large water reservoirs can be obtained. In a hydro-electric power station, water head is created by constructing a dam across a river or lake. From the dam, water is led to a water turbine. The water turbine captures the energy in the falling water and changes the hydraulic energy (*i.e.*, product of head and flow of water) into mechanical energy at the turbine shaft. The turbine drives the alternator which converts mechanical energy into electrical energy. Hydro-electric power stations are becoming very popular because the reserves of fuels (*i.e.*, coal and oil) are depleting day by day. They have the added importance for flood control, storage of water for irrigation and water for drinking purposes.

Advantages

(i) It requires no fuel as water is used for the generation of electrical energy.
(ii) It is quite neat and clean as no smoke or ash is produced.
(iii) It requires very small running charges because water is the source of energy which is avail- able free of cost.
(iv) It is comparatively simple in construction and requires less maintenance.
(v) It does not require a long starting time like a steam power station. In fact, such plants can be put into serviceinstantly.
(vi) It is robust and has a longer life.
(vii) Such plants serve many purposes. In addition to the generation of electrical energy, they also help in irrigation and controlling floods.
(viii) Although such plants require the attention of highly skilled persons at the time of construction, yet for operation, a few experienced persons may do the job well.

Disadvantages

(i) It involves high capital cost due to construction ofdam.
(ii) There is uncertainty about the availability of huge amount of water due to dependence on weather conditions.
(iii) Skilled and experienced hands are required to build theplant.
(iv) It requires highcost of transmission lines as the plant is located in hilly areas which are quite away from theconsumers.

Schematic Arrang ement of Hydroele ctric Power Station

Although a hydro-electric power station simply involves the conversion of hydraulic energy into electrical energy, yet it embraces many arrangements for proper working and efficiency. The schematic arrangement of a modern hydro-electric plant is shown in Fig. 2.2.

The dam is constructed across a river or lake and water from the catchment area collects at the back of the dam to form a reservoir. A pressure tunnel is taken off from the reservoir and water brought to the valve house at the start of the penstock. The valve house contains main sluice valves and automatic isolating valves. The former controls the water flow to the power house and the latter cuts off supply of water when the penstock bursts. From the valve house, water is taken to water turbine through a huge steel pipe known as *penstock*. The water turbine converts hydraulic energy into mechanical energy. The turbine drives the alternator which converts mechanical energy into electrical energy.

A surge tank (open from top) is built just before the valve house and protects the penstock from bursting in case the turbine gates suddenly close* due to electrical load being thrown off. When the gates close, there is a sudden stopping of water at the lower end of the penstock and consequently the penstock can burst like a paper log. The surge tank absorbs this pressure swing by increase in its level of water

Schematic arrangement of a Hydro-electric plant
Fig. 2.2

Constituents of Hydroelectric Plant

The constituents of a hydro-electric plant are

(1) hydraulic structures

(2) water turbines and electrical equipment.

Hydraulic structures. Hydraulic structures in a hydro-electric power station include dam, spillways, headworks, surge tank, penstock and accessory works.

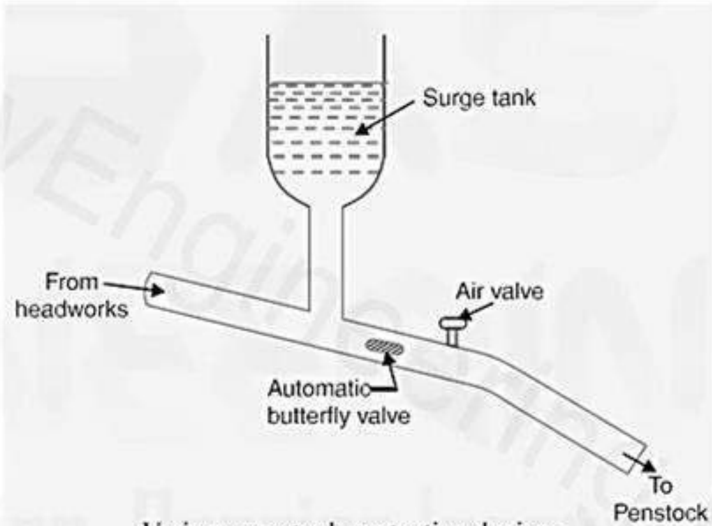
(i) ***Dam***. A dam is a barrier which stores water and creates water head. Dams are built of concrete or stone masonary, earth or rock fill. The type and arrangement depend upon the topography of the site. A masonary dam may be built in a narrow canyon. An earth dam may be best suited for a wide valley. The type of dam also depends upon the foundation conditions, localmaterials and transportation available, occurrenceof earthquakes and other hazards. At most of sites, more than one type of dam may be suitable and the one whichis most economical is chosen.

(ii) ***Spillways***. There are times when the river flow exceeds the storage capacity of the reservoir. Such a situation arises during heavy rainfall in the catchment area. In order to discharge the surplus water from the storage reservoir into the river on the down-stream side of the dam, spillways are used. Spillways are constructed of concrete piers on the top of the dam. Gates are provided between these piers and surplus water is discharged over the crest of thedam by opening these gates.

(iii) ***Headworks***. The headworks consists of the diversion structures at the head of an intake. They generally include booms and racks for diverting floating debris, sluices for by-passing debris and sediments and valves for controlling the flow of water to the turbine. The flow of water into and through headworks should be as smooth as possible to avoid head loss and cavitation. For this purpose, it is necessary to avoid sharp corners and abrupt contractions or enlargements.

(iv) *Surge tank.* Open conduits leading water to the turbine require no* protection. However, when closed conduits are used, protection becomes necessary to limit the abnormal pressure in the conduit. For this reason, closed conduits are always provided with a surge tank. A surge tank is a small reservoir or tank (open at the top) in which water level rises or falls to reduce the pressure swings in the conduit.

A surge tank is located near the beginning of the conduit.



Various penstock protective devices

Fig. 2.3

When the turbine is running at a steady load, there are no surges in the flow of water through the conduit *i.e.*, the quantity of water flowing in the conduit is just sufficient to meet the turbine requirements. However, when the load on the turbine decreases, the governor closes the gates of turbine, reducing water supply to the turbine. The excess water at the lower end of the conduit rushes back to the surge tank and increases its water level. Thus the conduit is prevented from bursting. On the other hand, when load on the turbine increases, additional water is drawn from the surge tank to meet the increased load requirement. Hence, a surge tank overcomes the abnormal pressure in the conduit when load on the turbine falls and acts as a reservoir during increase of load on the turbine.

(iv) ***Penstocks*.** Penstocks are open or closed conduits which carry water to the turbines. They are generally made of reinforced concrete or steel. Concrete penstocks are suitable for low heads (< 30 m) as greater pressure causes rapid deterioration of concrete. The steel pen- stocks can be designed for any head; the thickness of the penstock increases with the head or working pressure

Water turbines. Water turbines are used to convert the energy of falling water into mechanical energy. The principal types of water turbines are:

(i) Impulse turbines **(*ii*)** Reactionturbines

Nuclear Power Station:

A generating station in which nuclear energy is converted into electrical energy is known as a **nuclearpower station.**

In nuclear power station, heavy elements such as Uranium (U^{235}) or Thorium (Th^{232}) are subjected to nuclear fission in a special apparatus known as a *reactor*. The heat energy thus released is utilised in raising steam at high temperature and pressure. The steam runs the steam turbine which converts steam energy into mechanical energy. The turbine drives the alternator which converts mechanical energy into electricalenergy.

The most important feature of a nuclear power station is that huge amount of electrical energy can be produced from a relatively small amount of nuclear fuel as compared to other conventional types of power stations. It has been found that complete fission of 1 kg of Uranium (U^{235}) can produce as much energy as can be produced by the burning of 4,500 tons of high-grade coal. Although the recovery of principal nuclear fuels (*i.e.*, Uranium and Thorium) is difficult and expensive, yet the total energy content of the estimated world reserves of these fuels are considerably higher than those of conventional fuels, *viz*., coal, oil and gas. At present, energy crisis is gripping us and, therefore, nuclear energy can be successfully employed for producing low cost electrical energy on a large scale to meet the growing commercial and industrial demands.

Advantages

(i) The amount of fuel required is quite small. Therefore, there is a considerable saving inthe cost of fuel transportation.

(ii) A nuclear power plant requires less space as compared to any other type of the same size.

(iii) It has low running charges as a small amount of fuel is used for producing bulk electrical energy.

(iv) This type of plant is very economical for producing bulk electricpower.

(v) It can be located near the load centres because it does not require large quantities of water and need not be near coal mines. Therefore, the cost of primary distribution isreduced.

(vi) There are large deposits of nuclear fuels available all over the world. Therefore, such plants can ensure continued supply of electrical energy for thousands of years.

(vii) It ensures reliability of operation.

Disadvantages

(i) The fuel used is expensive and is difficult torecover.

(ii) The capital cost on a nuclear plant is very high as compared to other types of plants.

(iii) The erection and commissioning of the plant requires greater technical know-how.

(iv) The fission by-products are generally radioactive and may cause a dangerous amount of radioactive pollution.

(v) Maintenance charges are high due to lack of standardization. Moreover, high salaries of specially trained personnel employed to handle the plant further raise the cost.

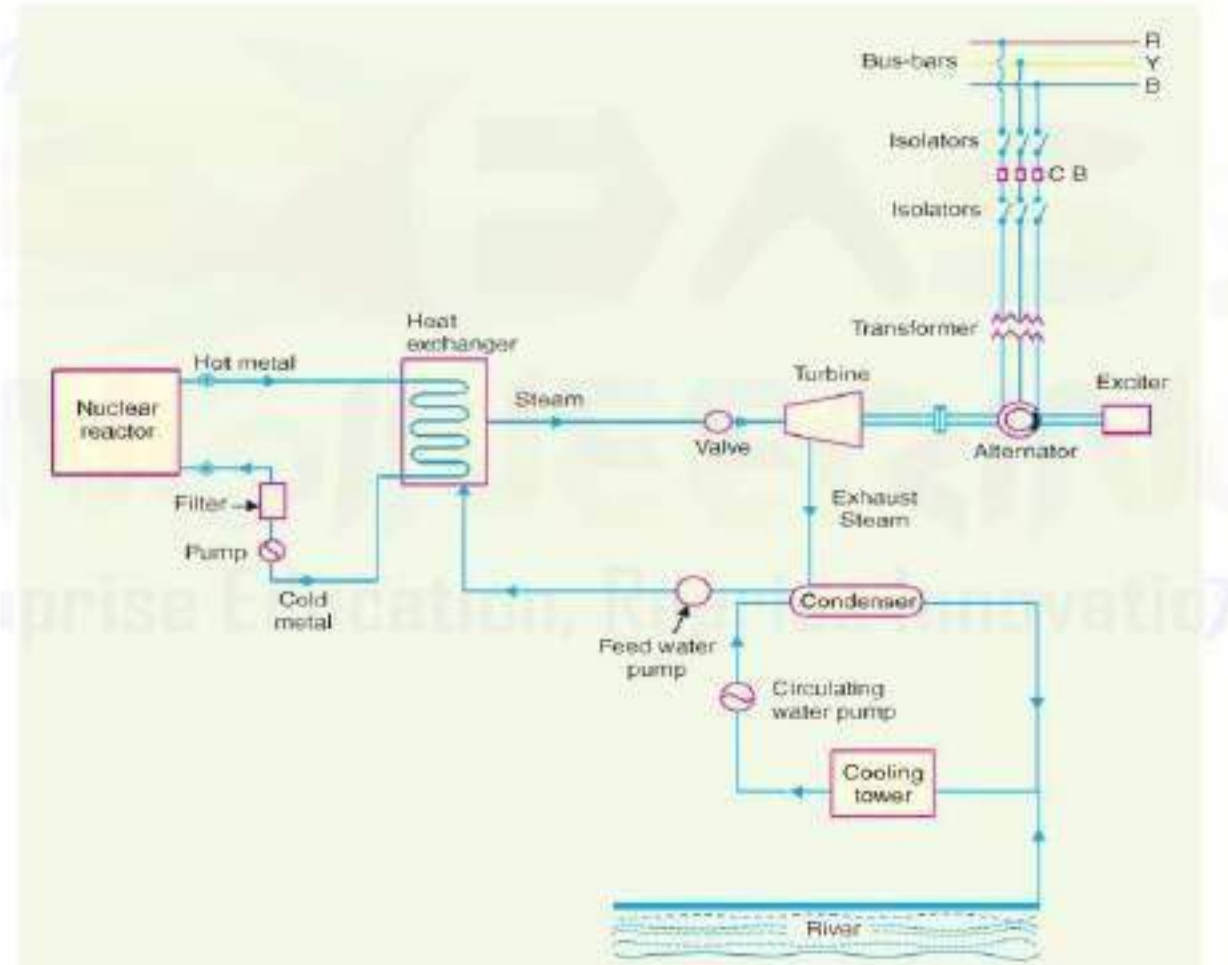
(vi) Nuclear power plants are not well suited for varying loads as the reactor does not respond to the load fluctuationsefficiently.

(vii) The disposal of the by-products, which are radioactive, is a big problem. They have either to be disposed of in a deep trench or in a sea away fromsea-shore.

Schematic Arrangement of Nuclear Power Station

The schematic arrangement of a nuclear power station is shown in Fig. 2.7. The whole arrangement can be divided into the following main stages :

(i) Nuclear reactor **(*ii*)** Heat exchanger **(*iii*)** Steam turbine **(*iv*)** Alternator.

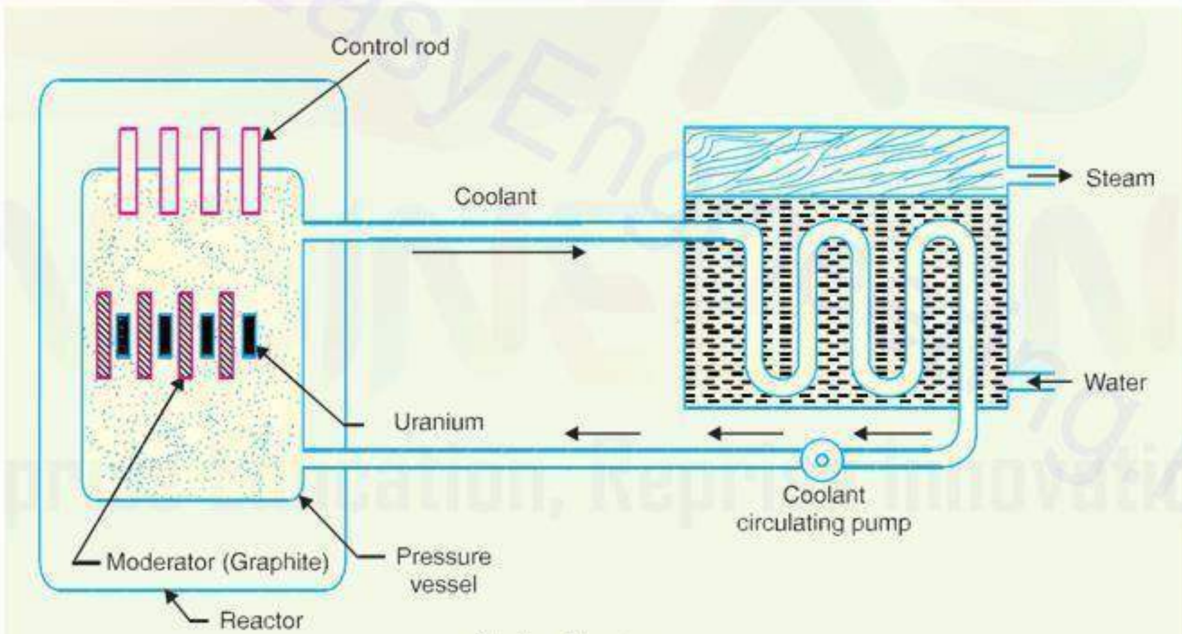


Schematic arrangement of Nuclear Power Station

Fig. 2.7

(i) **Nuclear reactor.** It is an apparatus in which nuclear fuel (U^{235}) is subjected to nuclear fission. It controls the *chain reaction** that starts once the fission is done. If the chain reaction is not controlled, the result will be an explosion due to the fast increase in the energy released.

A nuclear reactor is a cylindrical stout pressure vessel and houses fuel rods of Uranium, moderator and control rods (See Fig. 2.8). The fuel rods constitute the fission material and release huge amount of energy when bombarded with slow moving neutrons. The moderator consists of graphite rods which enclose the fuel rods. The moderator slows down the neutrons before they bombard the fuel rods. The control rods are of cadmium and are inserted into the reactor. Cadmium is strong neutron absorber and thus regulates the supply of neutrons for fission. When the control rods are pushed in deep enough, they absorb most of fission neutrons and hence few are available for chain reaction which, therefore, stops. However, as they are being withdrawn, more and more of these fission neutrons cause fission and hence the *intensity* of chain reaction (or heat produced) is increased. Therefore, by pulling out the control rods, power of the nuclear reactor is increased, whereas bypushing them in, it is reduced. In actual practice, the lowering or raising of control rods is accomplished automatically according to the requirement of load. The heat produced in thereactor is removed by the coolant, generally a sodium metal. The coolant carries the heat to the heat exchanger.



Nuclear Reactor
Fig. 2.8

(ii) **Heatexchanger.** The coolant gives up heat to the heat exchanger which is utilised in raising the steam. After giving up heat, the coolant is again fed to the reactor.

(iii) **Steamturbine.** The steam produced in the heatexchanger is led to the steam turbine through a valve. After doing a useful work in the turbine, the steam is exhausted to condenser. The condenser condenses the steam which is fed to the heat exchanger through feed water pump.

(iv) **Alternator.** The steam turbine drives the alternator which converts mechanical energy into electrical energy. The output from the alternator is delivered to the bus-bars through trans- former, circuit breakers and isolators.

Solar power plant:

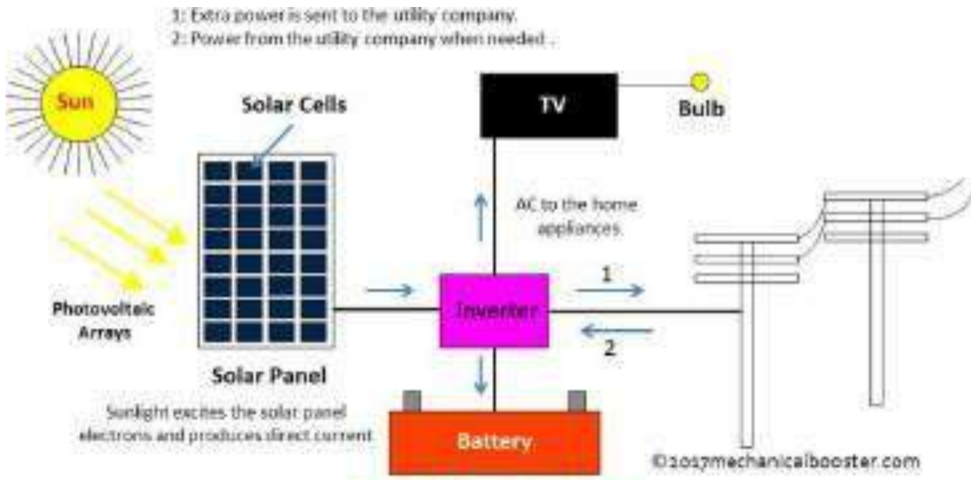
- **How does a Solar Panel Converts Sunlight into Electricity?**

Sunlight travels to earth in the form of small energy particles called photons. This photon strikes the p- type region and transfer its energy to hole and electron pair thus exciting the electron and it gets away from hole. The electric field we have due to potential difference at p-n junction makes its electron to travel to n-type region thus causing the current to flow.
But there a bit more to know, to make this electric field strong enough so that it must travel to n-type region and not recombine with the hole it has been separated from. To make this electric field strong the n-type and p-type regions are connected to negative and positive terminals of battery, this process is known as reverse bias condition. Doing this increases the probability of electron travelling all along the way to n-type region once separated from a hole. Thus increasing the efficiency of a solar panel.

- **Working Principle:**

The working principle is that we use the energy of photons to get the drift current flowing in the circuit using reversed bias p-n junction diode (p-type and n-type silicon combination).

- **Main Components:**



Working of Solar Power Plant

1. Solar Panels

It is the heart of the solar power plant. Solar panels consist a number of solar cells. We have got around 35 solar cells in one panel. The energy produced by each solar cell is very small, but combining the energy of 35 of them we have got enough energy to charge a 12-volt battery.

2. Solar Cells

It is the energy generating unit, made up of p-type and n-type silicon semiconductor. It's the heart of solar power plant.

3. Battery

Batteries are used to produce the power back or store the excess energy produced during day, to be supplied during night.

4. D.C. to A.C. Converter (Inverter)

Solar panels produce direct current which is required to be converted into alternating current to be supplied to homes or power grid.

Working of Solar Power Plant

As sunlight falls over a solar cell, a large number of photons strike the p-type region of silicon. Electron and hole pair will get separated after absorbing the energy of photon. The electron travels from p-type region to n-type region due to the action of electric field at p-n junction. Further the diode is reversed biased to increase this electric field. So, this current start flowing in the circuit for individual solar cell. We combine the current of all the solar cells of a solar panel, to get a significant output.

Solar power plant has a large number of solar panels connected to each other to get a large voltage output. The electrical energy coming from the combined effort of solar panels is stored in the Lithium ion batteries to be supplied at night time, when there is no sunlight.

Energy Storage
Storage of the energy generated by the solar panels is a important issue. Sometimes the unused energy generated during daytime is used to pump water to some height, so that it could be used to generate electricity using its potential energy when required or mainly at night time.

For current being Tesla is providing its industrial energy pack to store energy and currently it is lighting up an entire island. Tesla has also made an offer to Australia that it could provide its battery pack for emergency blackouts.

The cost of manufacturing of solar panels has decreased rapidly in last few years, same is said to be true with the industrial energy pack (Lithium ion batteries), as the production and demand increases their cost is going to decrease in coming few years.

Chapter2 Transmission of electric power

Layout of transmission and distribution scheme:

The conveyance of electric power from a power station to consumers' premises is known as **electric supply system**.

An electric supply system consists of three principal components *viz*., the power station, the transmission lines and the distribution system. Electric power is produced at the power stations which are located at favourable places, generally quite away from the consumers. It is then transmitted over large distances to load centres with the help of conductors known as transmission lines. Finally, it is distributed to a large number of small and big consumers through a distribution network

The electric supply system can be broadly classified into
(*i*)d.c. or a.c. system
(*ii*) overhead or underground system.
Now-a- days, 3-phase, 3-wire a.c. system is universally adopted for generation and transmission of electric power as an economical proposition. However, distribution of electric power is done by 3-phase, 4-wire a.c. system. The underground system is more expensive than the overhead system. Therefore, in our country, overhead system is mostly adopted for transmission and distribution of electric power.

Generating station: In Fig 7.1, G.S. represents the generating station where electric power is produced by 3-phase alternators operating in parallel. Theusual generation voltage is 11 kV. For economy in the transmission of electric power, the generation voltage (*i.e.*, 11 kV) is stepped upto 132 kV (or more) at the generating station with the help of 3-phase trans- formers. The transmission of electric power at high voltages has several advantages including the saving of conductor material and high transmission efficiency. It may appear advisable to use thehighest possible voltage for transmission of electric power to save conductor material and have other advantages. But there is a limit to which this voltage can be increased. It is because in- crease in transmission voltage introduces insulation problems as well as the cost of switchgear and transformer equipment is increased. Therefore, the choice of proper transmission voltage is essentially

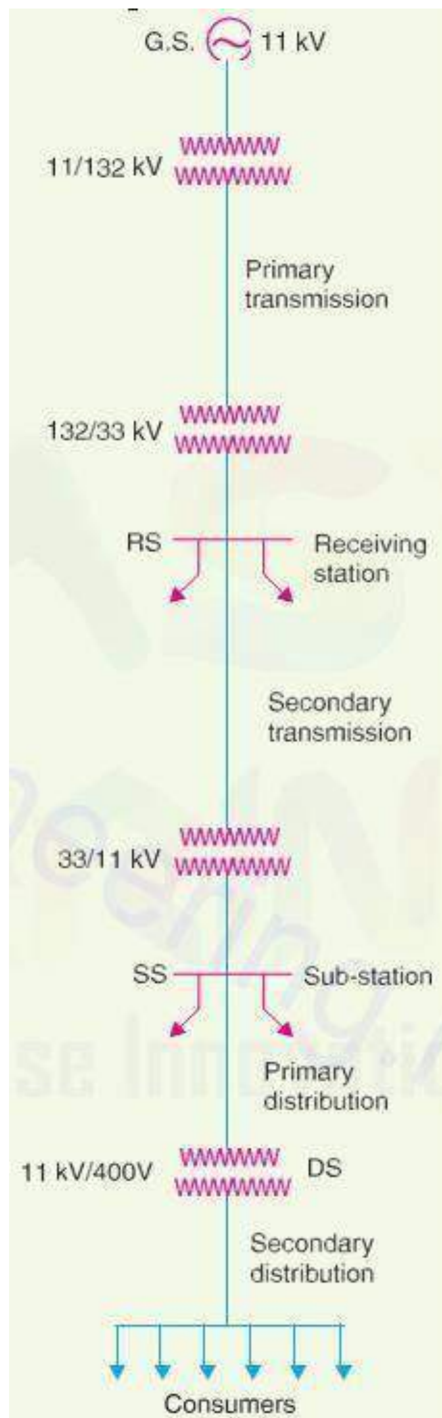


Fig. 7.1

a question of economics. Generally, the primary transmission is carried at 66 kV, 132 kV, 220 kV or 400 kV.

Primary transmission. The electric power at 132 kV is transmitted by 3-phase, 3-wire overhead system to the out- skirts of the city. This forms the primary transmission.

Secondary transmission. The primary transmission line terminates at the receiving station (*RS*) which usually lies at the outskirts of the city. At the receiving station, the voltage is re- duced to 33kV by step-down transformers. From this station, electric power is transmitted at 33kV by 3-phase, 3-wire over- head system to various sub-stations (*SS*) located at the strategic points in the city. This forms the secondary transmission.

Primary distribution. The secondary transmission line terminates at the sub-station (*SS*) where voltage is reduced from 33 kV to 11kV, 3-phase, 3-wire. The 11 kV lines run along the important road sides of the city. This forms the primary distribution. It may be noted that big con- sumers (having demand more than 50 kW) are generally supplied power at 11 kV for further handling with their ownsub-stations.

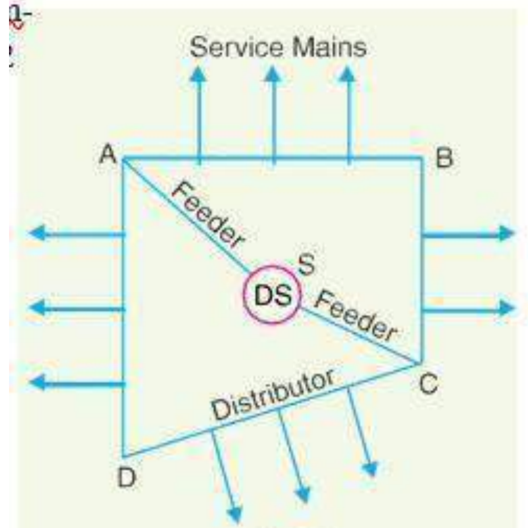


Fig. 7.2

secondary distribution.: The electric power from primary distribution line (11 kV) is delivered to distribution sub-stations (DS). These sub-stations are located near the consumers' localities and step down the voltage to 400 V, 3-phase, 4-wire for secondary distribution. The voltage between any two phases is 400 V and between any phase and neutral is 230 V. The single-phase residential lighting load is connected between any one phase and neutral, whereas 3-phase, 400 V motor load is connected across 3-phase lines directly.

It may be worthwhile to mention here thatsecondarydistributionsystemconsists of *feeders*, *distributors and service mains*. Fig. 7.2 shows the elements of low voltage distribution system. Feeders (*SC* or *SA*) radiating from the distributionsub-station (*DS*) supply power to the distributors (*AB*, *BC*, *CD* and *AD*). No consumer is given direct connection from the feeders. Instead, the consumers are connected to the distributors through their servicemains.

Voltage regulation and efficiency of transmission:

Voltage regulation. When a transmission line is carrying current, there is a voltage drop in the line due to resistance and inductance of the line. The result is that receiving end voltage (V_R) of the line is generally less than the sending end voltage (V_S). This voltage drop ($V_S - V_R$) in the line is expressed as a percentage of receiving end voltage V_R and is called voltage regulation.

The difference in voltage at the receiving end of a transmission line between conditions of no load and full load is called **voltage regulation** and is expressed as a percentage of the receiving end voltage.

Mathematically,

% age Voltage regulation = $(V_s - V_R) / V_R \times 100$

Obviously, it is desirable that the voltage regulation of a transmission line should be low *i.e.*, the increase in load current should make very little difference in the receiving end voltage.

Transmission efficiency. The power obtained at the receiving end of a transmission line is generally less than the sending end power due to losses in the line resistance.

The ratio of receiving end power to the sending end power of a transmission line is known as the **transmission efficiency** *of the line i.e.*

% age Transmission efficiency, η_T

$$= \frac{\text{Receiving end power}}{\text{Sending end power}} \times 100$$

$$= \left(V_R I_R \cos \phi_R / V_S I_S \cos \phi_S \right) \times 100$$

where V_R, I_R and $\cos \phi_R$ are the receiving end voltage, current and power factor while V_S, I_S and $\cos \phi_S$ are the corresponding values at the sending end..

Kelvin's law for economical size of conductor:

Economic Choice of Conductor Size:

The cost of conductor material is generally a very considerable part of the total cost of a transmission line. Therefore, the determination of proper size of conductor for the line is of vital importance. The most economical area of conductor is that for which the total annual cost of transmission line is minimum*. This is known as *Kelvin's Law* after Lord Kelvin who first stated it in 1881. The total annual cost of transmission line can be divided broadly into two parts *viz.*, annual charge oncapital outlay and annual cost of energy wasted in the conductor.

(i)Annual charge on capital outlay. This is on account of interest and depreciation on the capital cost of complete installation of transmission line. In case of overhead system, it will be the annual interest and depreciation on the capital cost of conductors, supports and insulators and the cost of their erection. Now, for an overhead line, insulator cost is constant, the conductor cost is proportional to the area of X-section and the cost of supports and their erection is partly constant and partly proportional to area of X-section of the conductor. Therefore, annual charge on an overhead transmission line can be expressed as :

Annual charge $= P_1 + P_2 a$

where P_1 and P_2 are constants and a is the area of X-section of the conductor.

(*ii*) *Annual cost of energy wasted.* This is on account of energy lost mainly‡ in the conductor due to I^2R losses. Assuming a constant current in the conductor throughout the year, the energy lost in the conductor is proportional to resistance. As resistance is inversely proportional to the area of X-section of the conductor, therefore, the energy lost in the conductor is inversely proportional to area of X-section. Thus, the annual cost of energy wasted in an overhead transmission line can be expressed as :

$$\text{Annual cost of energy wasted} = P_3/a \qquad ...(ii)$$

where P_3 is a constant.

$$\text{Total annual cost, } C = \text{exp. } (i) + \text{exp. } (ii)$$
$$= (P_1 + P_2\,a) + P_3/a$$
$$\therefore \qquad C = P_1 + P_2\,a + P_3/a \qquad ...(iii)$$

In exp. (*iii*), only area of X-section a is variable. Therefore, the total annual cost of transmission line will be minimum if differentiation of C w.r.t. a is zero *i.e.*

$$\frac{d}{da}(C) = 0$$

or $$\frac{d}{da}(P_1 + P_2\,a + P_3/a) = 0$$

or $$P_2 - \frac{P_3}{a^2} = 0$$

or $$P_2 = P_3/a^2$$

or $$P_2\,a = \frac{P_3}{a}$$

i.e. Variable part of annual charge = Annual cost of energy wasted

Therefore Kelvin's Law can also be stated in an another way *i.e. the most economical area of conductor is that for which the variable part* of annual charge is equal to the cost of energy losses per year.*

Graphical illustration of Kelvin's law. Kelvin's law can also be illustrated graphically by plotting annual cost against X-sectional area 'a' of the conductor as shown in Fig. 7.28. In the diagram, the straight line (1) shows the relation between the annual charge (*i.e.*, $P_1 + P_2a$) and the area of X-section a of the conductor. Similarly, the rectangular hyperbola (2) gives the relation between annual cost of energy wasted and X-sectional area a. By adding the ordinates of curves (1) and (2), the curve (3) is obtained. This latter curve shows the relation between total annual cost ($P_1 + P_2a + P_3/a$) of transmission line and area of X-section a. The lowest point on the curve (*i.e.*, point P) represents the most economical area of X-section.

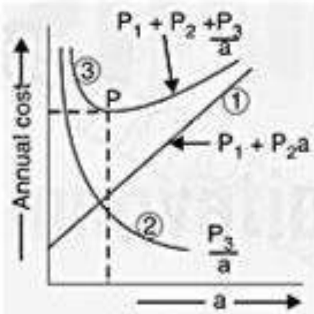


Fig. 7.28

Limitations of Kelvin's law. Although theoretically Kelvin's law holds good, there is often considerable difficulty in applying it to a proposed scheme of power transmission. In practice, the limitations of this law are :

(*i*) It is not easy to estimate the energy loss in the line without actual load curves, which are not available at the time of estimation.

(*ii*) The assumption that annual cost on account of interest and depreciation on the capital outlay is in the form $P_1 + P_2a$ is strictly speaking not true. For instance, in cables neither the cost of cable dielectric and sheath nor the cost of laying vary in this manner.

(iii) This law does not take into account several physical factors like safe current density, mechanical strength, corona loss etc.

(iv) The conductor size determined by this law may not always be practicable one because it may be too small for the safe carrying of necessary current.

(v) Interest and depreciation on the capital outlay cannot be determined accurately.

Corona and corona loss in transmission lines:

Corona

When an alternating potential difference is applied across two conductors whose spacing is large as compared to their diameters, there is no apparent change in the condition of atmospheric air sur- rounding the wires if the applied voltage is low. However, when the applied voltage exceeds a certain value, called *critical disruptive voltage,* the conductors are surrounded by a faint violet glowcalled corona.

The phenomenon of corona is accompanied by a hissing sound, production of ozone, power loss and radio interference. The higher the voltage is raised, the larger and higher the luminous envelope becomes, and greater are the sound, the power loss and the radio noise. If the applied voltage is increased to breakdown value, a flash-over will occur between the conductors due to the breakdown of air insulation.

The phenomenon of violet glow, hissing noise and production of ozone gas in an overhead transmission line is known as **corona.**

If the conductors are polished and smooth, the corona glow will be uniform throughout the length of the conductors, otherwise the rough points will appear brighter. With d.c. voltage, there is difference in the appearance of the two wires. The positive wire has uniform glow about it, while the negative conductor has spottyglow.

Theory of corona formation.

Someionisation isalwayspresent in air due to cosmic rays, ultra- violet radiations andradioactivity. Therefore, under normal conditions, the air around the conductors contains some ionised particles (*i.e.,* free electrons and +ve ions) and neutral molecules. When p.d. is applied between the conductors, potential gradient is set up in the air which will have maximum value at the conductor surfaces. Under the influence of potential gradient, the existing free electrons acquire greater velocities. The greater the applied voltage, the greater the potential gradient and more is the velocity of free electrons.

When the potential gradient at the conductor surface reaches about 30 kV per cm (max. value), the velocity acquired by the free electrons is sufficient to strike a neutral molecule with enough force to dislodge one or more electrons from it. This produces another ion and one or more free electrons, which is turn are accelerated until they collide with other neutral molecules, thus producing other ions. Thus, the process of ionisation is cummulative. The result of this ionisation is that either corona is formed or spark takes place between the conductors.

Factors Affecting Corona

The phenomenon of corona is affected by the physical state of the atmosphere as well as by the conditions of the line. The following are the factors upon which corona depends:

1.***Atmosphere***. As corona is formed due to ionization of air surrounding the conductors, there-fore, it is affected by the physical state of atmosphere. In the stormy weather, the number of ions is more than normal and as such corona occurs at much less voltage as compared with fair weather.

2.***Conductor size***. The corona effect depends upon the shape and conditions of the conductors. The rough and irregular surface will give rise to more corona because unevenness of the surface decreases the value of breakdown voltage. Thus a stranded conductor has ir- regular surface and hence gives rise to more corona that a solid conductor.

3. ***Spacing between conductors***. If the spacing between the conductors is made very large as compared to their diameters, there may not be any corona effect. It is because larger dis- tance between conductors reduces the electro-static stresses at the conductor surface, thus avoiding corona formation.

4. ***Line voltage.*** The line voltage greatly affects corona. If it is low, there is no change in the condition of air surrounding the conductors and hence no corona is formed. However, if the line voltage has such a value that electrostatic stresses developed at the conductor surface make the air around the conductor conducting, then corona is formed

8.12 Important Terms

The phenomenon of corona plays an important role in the design of an overhead transmission line. Therefore, it is profitable to consider the following terms much used in the analysis of corona effects:

(*i*) **Critical disruptive voltage.** *It is the minimum phase-neutral voltage at which corona occurs.*

Consider two conductors of radii r cm and spaced d cm apart. If V is the phase-neutral potential, then potential gradient at the conductor surface is given by:

$$g = \frac{V}{r \log_e \frac{d}{r}} \text{ volts / cm}$$

In order that corona is formed, the value of g must be made equal to the breakdown strength of air. The breakdown strength of air at 76 cm pressure and temperature of 25°C is 30 kV/cm (*max*) or

21·2 kV/cm (*r.m.s.*) and is denoted by g_o. If V_c is the phase-neutral potential required under these conditions, then,

$$g_o = \frac{V_c}{r \log_e \frac{d}{r}}$$

where g_o = breakdown strength of air at 76 cm of mercury and 25°C
= 30 kV/cm (*max*) or 21·2 kV/cm (*r.m.s.*)

$\therefore$ Critical disruptive voltage, $V_c = g_o r \log_e \frac{d}{r}$

The above expression for disruptive voltage is under standard conditions *i.e.*, at 76 cm of Hg and 25°C. However, if these conditions vary, the air density also changes, thus altering the value of g_o. The value of g_o is directly proportional to air density. Thus the breakdown strength of air at a barometric pressure of b cm of mercury and temperature of t°C becomes δg_o where

$$\delta = \text{air density factor} = \frac{3 \cdot 92 b}{273 + t}$$

Under standard conditions, the value of $\delta = 1$.

$\therefore$ Critical disruptive voltage, $V_c = g_o \delta r \log_e \frac{d}{r}$

Correction must also be made for the surface condition of the conductor. This is accounted for by multiplying the above expression by irregularity factor m_o.

$\therefore$ Critical disruptive voltage, $V_c = m_o g_o \delta r \log_e \frac{d}{r}$ kV/phase

where m_o = 1 for polished conductors
= 0·98 to 0·92 for dirty conductors
= 0·87 to 0·8 for stranded conductors

(*ii*) **Visual critical voltage.** *It is the minimum phase-neutral voltage at which corona glow appears all along the line conductors.*

It has been seen that in case of parallel conductors, the corona glow does not begin at the disruptive voltage V_c but at a higher voltage V_v, called *visual critical voltage*. The phase-neutral effective value of visual critical voltage is given by the following empirical formula :

$$V_v = m_v g_o \delta r \left(1 + \frac{0 \cdot 3}{\sqrt{\delta r}}\right) \log_e \frac{d}{r} \text{ kV/phase}$$

where m_v is another irregularity factor having a value of 1·0 for polished conductors and 0·72 to 0·82 for rough conductors.

(*iii*) **Power loss due to corona.** Formation of corona is always accompanied by energy loss which is dissipated in the form of light, heat, sound and chemical action. When disruptive voltage is exceeded, the power loss due to corona is given by :

$$P = 242 \cdot 2 \left(\frac{f + 25}{\delta}\right) \sqrt{\frac{r}{d}} \left(V - V_c\right)^2 \times 10^{-5} \text{ kW / km / phase}$$

where f = supply frequency in Hz
V = phase-neutral voltage (*r.m.s.*)
V_c = disruptive voltage (*r.m.s.*) per phase

Factors Affecting Corona

The phenomenon of corona is affected by the physical state of the atmosphere as well as by the conditions of the line. The following are the factors upon which corona depends :

- *Atmosphere.* As corona is formed due to ionisation of air surrounding the conductors, there- fore, it is affected by the physical state of atmosphere. In the stormy weather, the number of ions is more than normal and as such corona occurs at much less voltage as compared with fair weather.
- *Conductor size.* The corona effect depends upon the shape and conditions of the conductors. The rough and irregular surface will give rise to more corona because unevenness of the surface decreases the value of breakdown voltage. Thus a stranded conductor has ir- regular surface and hence gives rise to more corona that a solid conductor.
- *Spacing between conductors.* If the spacing between the conductors is made very large as compared to their diameters, there may not be any corona effect. It is because larger distance between conductors reduces the electro-static stresses at the conductor surface, thus avoiding corona formation.
- *Line voltage.* The line voltage greatly affects corona. If it is low, there is no change in the condition of air surrounding the conductors and hence no corona is formed. However, if the line voltage has such a value that electrostatic stresses developed at the conductor surface make the air around the conductor conducting, then corona is formed.

Advantages and Disadvantages of Corona

Corona has many advantages and disadvantages. In the correct design of a high voltage overhead line, a balance should be struck between the advantages and disadvantages.

Advantages

- Due to corona formation, the air surrounding the conductor becomes conducting and hence virtual diameter of the conductor is increased. The increased diameter reduces the electro- static stresses between theconductors.
- Corona reduces the effects of transients produced bysurges.

Disadvantages

- Corona is accompanied by a loss of energy. This affects the transmission efficiency of the line.
- Ozone is produced by corona and may cause corrosion of the conductor due to chemical action.
- The current drawn by the line due to corona is nonsinusoidal and hence nonsinusoidal voltage drop occurs in the line. This may cause inductive interference with neighbouring communication lines.

Methods of Reducing Corona Effect

It has been seen that intense corona effects are observed at a working voltage of 33 kV or above. Therefore, careful design should be made to avoid corona on the sub-stations or bus-bars rated for 33 kV and higher voltages otherwise highly ionised air may cause flash-over in the insulators or between the phases, causing considerable damage to the equipment. The corona effects can be reduced by the following methods:

- ***By increasing conductor size.*** By increasing conductor size, the voltage at which corona occurs is raised and hence corona effects are considerably reduced. This is one of the reasons that *ACSR* conductors which have a larger cross-sectional area are used in transmission lines.

- ***By increasing conductor spacing.*** By increasing the spacing between conductors, the volt- age at which corona occurs is raised and hence corona effects can be eliminated. However, spacing cannot be increased too much otherwise the cost of supporting structure (*e.g.,* big- ger cross arms and supports) may increase to a considerableextent.

Example *A 3-phase line has conductors 2 cm in diameter spaced equilaterally 1 m apart. If the dielectric strength of air is 30 kV (max) per cm, find the disruptive critical voltage for the line. Take air density factor $\delta = 0{\cdot}952$ and irregularity factor $m_o = 0{\cdot}9$.*

Solution.

Conductor radius, $r = 2/2 = 1$ cm

Conductor spacing, $d = 1$ m $= 100$ cm

Dielectric strength of air, $g_o = 30$ kV/cm (*max.*) $= 21{\cdot}2$ kV (*r.m.s.*) per cm

Disruptive critical voltage, $V_c = m_o\, g_o\, \delta\, r \log_e (d/r)$ kV*/phase (*r.m.s.* value)

$= 0{\cdot}9 \times 21{\cdot}2 \times 0{\cdot}952 \times 1 \times \log_e 100/1 = 83{\cdot}64$ kV/phase

$\therefore$ Line voltage (*r.m.s.*) $= \sqrt{3} \times 83{\cdot}64 = 144{\cdot}8$ kV

Example *A 132 kV line with 1·956 cm dia. conductors is built so that corona takes place if the line voltage exceeds 210 kV (r.m.s.). If the value of potential gradient at which ionisation occurs can be taken as 30 kV per cm, find the spacing between the conductors.*

Solution.

Assume the line is 3-phase.

Conductor radius, $r = 1\cdot956/2 = 0\cdot978$ cm

Dielectric strength of air, $g_o = 30/\sqrt{2} = 21\cdot2$ kV (*r.m.s.*) per cm

Disruptive voltage/phase, $V_c = 210/\sqrt{3} = 121\cdot25$ kV

Assume smooth conductors (*i.e.*, irregularity factor $m_o = 1$) and standard pressure and temperature for which air density factor $\delta = 1$. Let d cm be the spacing between the conductors.

$\therefore$ Disruptive voltage (*r.m.s.*) per phase is

$$V_c = m_o g_o \delta r \log_e (d/r) \text{ kV}$$
$$= 1 \times 21\cdot2 \times 1 \times 0\cdot978 \times \log_e (d/r)$$

or $$121\cdot25 = 20\cdot733 \log_e (d/r)$$

or $$\log_e \frac{d}{r} = \frac{121\cdot25}{20\cdot733} = 5\cdot848$$

or $$2\cdot3 \log_{10} d/r = 5\cdot848$$

or $$\log_{10} d/r = 5\cdot848/2\cdot3 = 2\cdot5426$$

or $$d/r = \text{Antilog } 2\cdot5426$$

or $$d/r = 348\cdot8$$

$\therefore$ Conductor spacing, $d = 348\cdot8 \times r = 348\cdot8 \times 0\cdot978 = $ **341 cm**

Example *A 3-phase, 220 kV, 50 Hz transmission line consists of 1·5 cm radius conductor spaced 2 metres apart in equilateral triangular formation. If the temperature is 40°C and atmospheric pressure is 76 cm, calculate the corona loss per km of the line. Take $m_o = 0{\cdot}85$.*

Solution.

As seen from Art. 8.12, the corona loss is given by :

$$P = \frac{242\cdot2}{\delta}(f+25)\sqrt{\frac{r}{d}}(V - V_c)^2 \times 10^{-5} \text{ kW/km/phase}$$

Now, $$\delta = \frac{3\cdot92\, b}{273+t} = \frac{3\cdot92 \times 76}{273+40} = 0\cdot952$$

Assuming $$g_o = 21\cdot2 \text{ kV/cm } (r.m.s.)$$

$\therefore$ Critical disruptive voltage per phase is

$$V_c = m_o g_o \delta r \log_e d/r \text{ kV}$$
$$= 0\cdot85 \times 21\cdot2 \times 0\cdot952 \times 1\cdot5 \times \log_e 200/1\cdot5 = 125\cdot9 \text{ kV}$$

Supply voltage per phase, $V = 220/\sqrt{3} = 127$ kV

Substituting the above values, we have corona loss as:

$$P = \frac{242\cdot2}{0\cdot952}(50+25) \times \sqrt{\frac{1\cdot5}{200}} \times (127 - 125\cdot9)^2 \times 10^{-5} \text{ kW/phase/km}$$
$$= \frac{242\cdot2}{0\cdot952} \times 75 \times 0\cdot0866 \times 1\cdot21 \times 10^{-5} \text{ kW/km/phase}$$
$$= 0\cdot01999 \text{ kW/km/phase}$$

$\therefore$ Total corona loss per km for three phases

$$= 3 \times 0\cdot01999 \text{ kW} = \mathbf{0\cdot05998 \text{ kW}}$$

Example *A certain 3-phase equilateral transmission line has a total corona loss of 53 kW at 106 kV and a loss of 98 kW at 110·9 kV. What is the disruptive critical voltage? What is the corona loss at 113 kV?*

Solution.

The power loss due to corona for 3 phases is given by :

$$P = 3 \times \frac{242 \cdot 2\,(f+25)}{\delta} \sqrt{\frac{r}{d}}\,(V - V_c)^2 \times 10^{-5} \text{ kW/km}$$

As f, δ, r and d are the same for the two cases,

$$\therefore \qquad P \propto (V - V_c)^2$$

For first case, $P = 53$ kW and $V = 106/\sqrt{3} = 61 \cdot 2$ kV

For second case, $P = 98$ kW and $V = 110 \cdot 9/\sqrt{3} = 64$ kV

$$\therefore \qquad 53 \propto (61 \cdot 2 - V_c)^2 \qquad \text{...}(i)$$

and

$$98 \propto (64 - V_c)^2 \qquad \text{...}(ii)$$

Dividing [(*ii*)/(*i*)], we get,

$$\frac{98}{53} = \frac{(64 - V_c)^2}{(61 \cdot 2 - V_c)^2}$$

or

$$V_c = 54 \text{ kV}$$

Let W kilowatt be the power loss at 113 kV.

$$\therefore \qquad W \propto \left(\frac{113}{\sqrt{3}} - V_c\right)^2$$

$$\propto (65 \cdot 2 - 54)^2 \qquad \text{...}(iii)$$

Dividing [(*iii*)/(*i*)], we get,

$$\frac{W}{53} = \frac{(65 \cdot 2 - 54)^2}{(61 \cdot 2 - 54)^2}$$

$$\therefore \qquad W = (11 \cdot 2/7 \cdot 2)^2 \times 53 = 128 \text{ kW}$$

TUTORIAL PROBLEMS

1. Estimate the corona loss for a three-phase, 110 kV, 50 Hz, 150 km long transmission line consisting of three conductors each of 10 mm diameter and spaced 2·5 m apart in an equilateral triangle formation. The temperature of air is 30°C and the atmospheric pressure is 750 mm of mercury. Take irregularity factor as 0·85. Ionisation of air may be assumed to take place at a maximum voltage gradient of 30 kV/cm. **[316·8 kW]**
2. Taking the dielectric strength of air to be 30 kV/cm, calculate the disruptive critical voltage for a 3-phase line with conductors of 1 cm radius and spaced symmetrically 4 m apart. **[220 kV line voltage]**
3. A 3-phase, 220 kV, 50 Hz transmission line consists of 1·2 cm radius conductors spaced 2 m at the corners of an equilateral triangle. Calculate the corona loss per km of the line. The condition of the wire is smoothly weathered and the weather is fair with temperature of 20°C and barometric pressure of 72·2 cm of Hg. **[2·148 kW]**

LECTURE NOTES

ON

GENERATION, TRANSMISSION DISTRIBUTION (Part II)

BRANCH- ELECTRICAL ENGINEERING

4th Semester

Department of Electrical Engineering

S. K. D. A. V. GOVERNMENT POLYTECHNIC

ROURKELA 12

CHAPTER3

OVERHEAD LINES:

Types of supports, size and spacing of conductor:

Main Components of Overhead Lines:

An overhead line may be used to transmit or distribute electric power. The succcessful operation of an overhead line depends to a great extent upon the mechanical design of the line. While constructing an overhead line, it should be ensured that mechanical strength of the line is such so as to provide against the most *probable* weather conditions. In general, the main components of an overhead lineare:

- ***Conductors*** which carry electric power from the sending end station to the receiving end station.
- ***Supports*** whichmay be poles or towers and keep the conductors at asuitable level above theground.
- ***Insulators*** which are attached to supports and insulate the conductors from the ground.
- ***Cross arms*** which provide support to theinsulators.
- ***Miscellaneous items*** such as phase plates, danger plates, lightning arrestors, anti-climbing wires etc.

Conductor Materials:

The conductor is one of the important items as most of the capital outlay is invested for it. Therefore, proper choice of material and size of the conductor is of considerable importance.The conductor material used for transmission and distribution of electric power should have thefollowing properties :

- high electricalconductivity.
- high tensile strength in order to withstand mechanical stresses.
- low cost so that it can be used for longdistances.
- low specific gravity so that weight per unit volume is small.

All above requirements are not found in a single material. Therefore, while selecting a conductor material for a particular case, a compromise is made between the cost and the required electrical and mechanical properties.

Commonly used conductor materials. The most commonly used conductor materials for over- head lines are *copper, aluminium, steel-cored aluminium, galvanised steel* and *cadmium copper*. The choice of a particular material will depend upon the cost, the required electrical and mechanical properties and the local conditions.

1. Copper. Copper is an ideal material for overhead lines owing to its high electrical conductivity and greater tensile strength. It is always used in the hard-drawn form as stranded conductor. Copper has high current density *i.e.*, the current carrying capacity of copper per unit of X-sec- tional area is quite large. This leads to two advantages. Firstly, smaller X-sectional area of conductor is required and secondly, the area offered by the conductor to wind loads is reduced.

Moreover, this metal is quite homogeneous, durable and has high scrapvalue.

2. Aluminium. Aluminium is cheap and light as compared to copper but it has much smaller conductivity and tensile strength. The relative comparison of the two materials is briefed below: The conductivity of aluminium is 60% that of copper. The smaller conductivity of aluminium means that for any particular transmission efficiency, the X-sectional area of conductor must be larger in aluminium than in copper. For the same resistance, the diameter of aluminium conductor is about 1·26 times the diameter of copperconductor.

- The specific gravity of aluminium (2·71 gm/cc) is lower than that of copper (8·9 gm/cc). Therefore, an aluminium conductor has almost one-half the weight of equivalent copper conductor. For this reason, the supporting structures for aluminium need not be made so strong as that of copper conductor.
- Aluminiumconductor being light, isliable togreater swings and hence largercross-arms are required.
- Due to lower tensile strength and higher co-efficient of linear expansion of aluminium, the sag is greater in aluminiumconductors.

3. Steel cored aluminium. Due to low tensile strength, aluminium conductors produce greater sag. This prohibits their use for larger spans and makes them unsuitable for long distance transmis- sion. In order to increase the tensile strength, the aluminium conductor is reinforced with a coreofgalvanised steel wires. The compositeconductor thus obtained is known as *steel cored aluminium* and is abbreviated as A.C.S.R. (aluminium conductor steel reinforced).

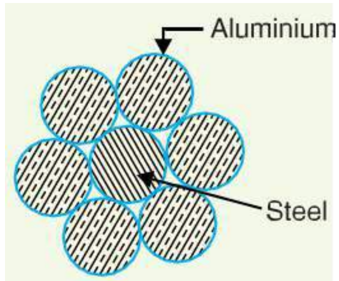


Fig. 8.1

Steel-cored aluminium conductor consists of central core of galvanised steel wires surrounded by a number of aluminium strands. Usually, diam- eter of both steel and aluminium wires is the same. The X-section of the two metals are generally in the ratio of 1 : 6 but can be modified to 1 : 4 in order to get more tensile strength for the conductor. Fig. 8.1 shows steel cored aluminium conductor having one steel wire surrounded by six wires of aluminium. The result of this composite conductor is that steel core takes greater percentage of mechanical strength while aluminium strands carry the bulk of current. The steel cored aluminium conductors have the following advantages :

- The reinforcement with steel increases the tensile strength but at the same time keeps the compositeconductor light. Therefore, steel coredaluminium conductorswill producesmaller sag and hence longer spans can be used.
- Due to smaller sag with steel cored aluminium conductors, towers of smaller heights can be used.

4. Galvanised steel. Steel has very high tensile strength. Therefore, galvanised steelconductors can be used for extremely long spans or for short line sections exposed to abnormally high stresses due to climatic conditions. They have been found very suitable in rural areas where cheap- ness is the main consideration. Due to poor conductivity and high resistance of steel, such conductors are not suitable for transmitting large power over a long distance. However, they can be used to advantage for transmitting a small power over a small distance where the size of the copper conductor desirable from economic considerations would be too small and thus unsuitable for use because of poor mechanical strength.

5. Cadmium copper. The conductor material now being employed in certain cases is copper alloyed withcadmium. An addition of 1% or 2% cadmium to copper increases the tensile strength by about 50% and the conductivity is only reduced by 15% below that of pure copper. Therefore, cadmium copper conductor can be useful for exceptionally long spans. However, due to highcost of cadmium, such conductors will be economical only for lines of small X-section *i.e.*, where the cost of conductor material is comparatively small compared with the cost of supports.

Line Supports:

The supporting structures for overhead line conductors are various types of poles and towers called *line supports*. In general, the line supports should have the following properties :

- High mechanical strength to withstand the weight of conductors and wind loads etc.
- Light in weight without the loss of mechanical strength.
- Cheap in cost and economical tomaintain.
- Longer life.
- Easy accessibility of conductors formaintenance.

The line supports used for transmission and distribution of electric power are of various types including *wooden poles, steel poles, R.C.C. poles* and *lattice steel towers*. The choice of supporting structure for a particular case depends upon the line span, X-sectional area, line voltage, cost and local conditions.

1. **Wooden poles.** These are made of seasoned wood (sal or chir) and are suitable for lines of moderate X-sectional area and of relatively shorter spans, say upto 50 metres. Such supports are cheap, easily available, provide insulating properties and, therefore, are widely used for distirbution purposes in rural areas as an economical proposition. The wooden poles generally tend to rot below the ground level, causing foundation failure. In order to prevent this, the portion of the pole belowthe ground level is impregnated with preservative compounds like *creosote oil*.

The main objections to wooden supports are : **(*i*)** tendency to rot below the ground level **(*ii*)** comparatively smaller life (20-25 years) **(*iii*)** cannot be used for voltages higher than 20 kV **(*iv*)** less mechanical strength and **(*v*)** require periodical inspection.

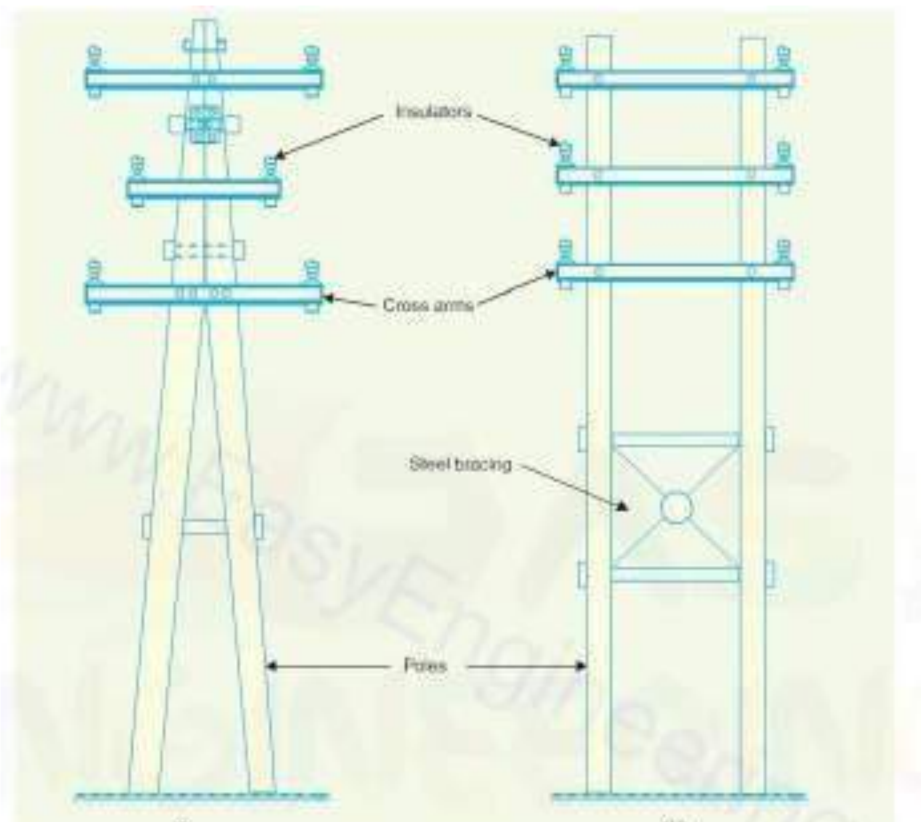
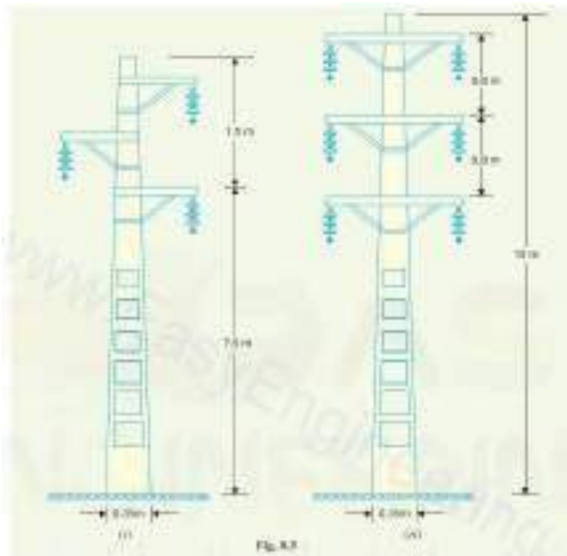


Fig. 8.2 Wooden poles

2. Steel poles. The steel poles are often used as a substitute for wooden poles. They possess greater mechanical strength, longer life and permit longer spans to be used. Such poles are generally used for distribution purposes in the cities. This type of supports need to be galvanised or painted in order to prolong its life.

3. RCC poles. The reinforced concrete poles have become very popular as line supports in recent years. They have greater mechanical strength, longer life and permit longer spans than steel poles. Moreover, they give good outlook, require little maintenance and have good insulating properties. Fig. 8.3 shows R.C.C. poles for single and double circuit. The holes in the poles facilitate the climbing of poles and at the same time reduce the weight of line supports.

The main difficulty with the use of these poles is the high cost of transport owing to their heavy weight. Therefore, such poles are often manufactured at the site in order to avoid heavy cost of transportation.



4. Steeltowers. In practice, wooden, steel and reinforced concrete poles are used for distribution purposes at low voltages, say upto 11 kV. However, for long distance transmission at higher voltage, steel towers are invariably employed. Steel towers have greater mechanical strength, longer life, can withstand most severe climatic conditions and permit the use of longer spans. The risk of interrupted service due to broken or punctured insulation is considerably reduced owing to longer spans. Tower footings are usually grounded by driving rods into the earth. This minimises the lightning troubles as each tower acts as a lightning conductor.

Insulators:

The overhead line conductors should be supported on the poles or towers in such a way that currents from conductorsdo notflow to earth through supports *i.e.*, lineconductorsmustbeproperly insulated from supports. This is achieved by securing line conductors to supports with the help of *insulators*. The insulators provide necessary insulation between line conductors and supports and thus prevent any leakage current from conductors to earth. In general, the insulators should have the following desirable properties :

- High mechanical strength in order to withstand conductor load, wind load etc.
- High electrical resistance of insulator material in order to avoid leakage currents to earth.
- High relative permittivity of insulator material in order that dielectric strength is high.
- The insulator material should be non-porous, free from impurities and cracks otherwise the permittivity will belowered.
- High ratio of puncture strength toflashover.

Types of Insulators:

The successful operation of an overhead line depends to a considerable extent upon the proper selection of insulators. There are several types of insulators but the most commonly used are pin type, suspension type, strain insulator and shackleinsulator.

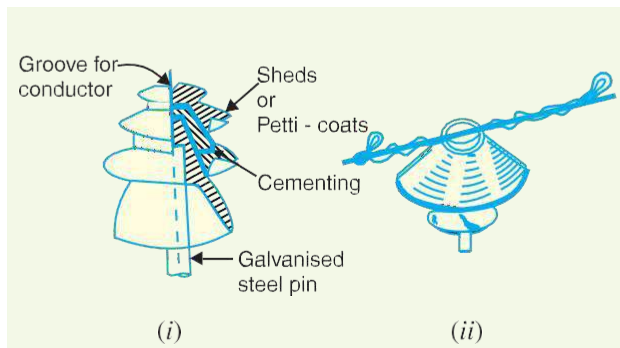


Fig. 8.5. Pin-type insulator

1. Pin type insulators. The part section of a pin type insulator is shown. As the name suggests, the pin type insulator is secured to the cross-arm on the pole. There is a groove on the upper end of the insulator for housing the conductor. The conductor passes through this groove and is bound by the annealed wire of the same material as the conductor.

Pin type insulators are used for transmission and distribution of electric power at voltages upto 33 kV. Beyond operating voltage of 33 kV, the pin type insulators become too bulky and hence uneconomical.

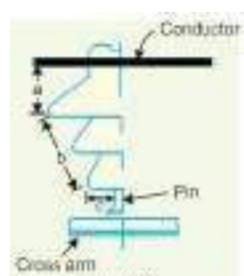


Fig. 8.6

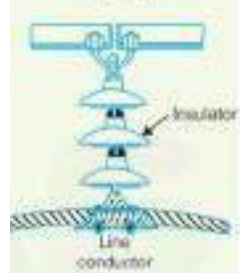


Fig. 8.7

Causes of insulator failure.

Insulators are required to withstand both mechanical and electrical stresses. The latter type is pirmarily due to line voltageand maycause the breakdown of the insulator. The electrical break- down of the insulator can occur either by *flash-over* or *puncture.* In flash- over, an arcoccurs between theline conductor andinsulator pin (*i.e.*, earth) and the discharge jumps across the *air gaps, following shortest dis- tance. Fig. 8.6 shows the arcing distance (*i.e.* $a + b + c$) for the insula- tor. In case of flash-over, the insulator will continue to act in its proper capacity unless extreme heat produced by the arc destroys the insulator.

In case of puncture, the discharge occurs from conductor to pin through the body of the insulator. When such breakdown is involved, the insulator is permanently destroyed due to excessive heat. In practice, sufficient thickness of porcelain is provided in the insulator to avoid puncture by the line voltage. The ratio of puncture strength to flash- over voltage is known as safety factor i.e.

$$\text{Safety factor of insulator} = \frac{\text{Puncture strength}}{\text{Flash - over voltage}}$$

2 Suspension type insulators. The cost of pin type insulator increases rapidly as the working voltage is increased. Therefore, this type of insulator is not economical beyond 33 kV. For high voltages (>33 kV), it is a usual practice to use suspension type insulators. It consist of a number of porcelain discs connected in series by metal links in the form of a string. The conductor is suspended at the bottom end of this string while the other end of the string is secured to the cross-arm of the tower. Each unit or disc is designed for low voltage, say 11 kV. The numberof discs in series would obviously depend upon the working voltage. For instance, if the working volt- age is 66 kV, then six discs in series will be provided on thestring.

Advantages

- Suspension type insulators are cheaper than pin type insulators for voltages beyond 33 kV.
- Each unit or disc of suspension type insulator is designed for low voltage, usually 11 kV. Depending upon the working voltage, the desired number of discs can be connected in series.
- If anyone disc is damaged, the whole string does not become useless because the damaged disc can be replaced by the sound one.
- The suspension arrangement provides greater flexibility to the line. The connection at the cross arm is such that insulator string is free to swing in any direction and can take up the position where mechanical stresses are minimum.
- In case of increased demand on the transmission line, it is found more satisfactory to supply the greater demand by raising the line voltage than to provide another set of conductors. The additional insulation required for the raised voltage can be easily obtained in the sus- pension arrangement by adding the desired number ofdiscs.
- The suspension type insulators are generally used with steel towers. As the conductors run below the earthed cross-arm of the tower, therefore, this arrangement provides partial protection from lightning.

3. Strain insulators. When there is a dead end of the line or there is corner or sharp curve, the line is subjected to greater tension. In order to relieve the line of excessive tension, strain insulators are used. For low voltage lines (< 11 kV), shackle insulators are used as strain insulators. However, for high voltage transmission lines, strain insulator consists of an assembly of suspension insulators as shown in Fig. 8.8. The discs of strain insulators are used in the vertical plane. When the tension in lines is exceedingly high, as at long river spans, two or more strings are used in parallel.

4. Shackle insulators. In early days, the shackle insulators were used as strain insulators. But now a days, they are frequently used for low voltage distribution lines. Such insulators can be used either in a horizontal position or in a vertical position. They can be directly fixed to the pole with a bolt or to the cross arm. Fig. 8.9 shows a shackle insulator fixed to the pole. The conductor in the groove is fixed with a soft bindingwire.

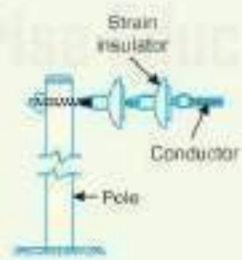


Fig. 8.8. Strain insulator.

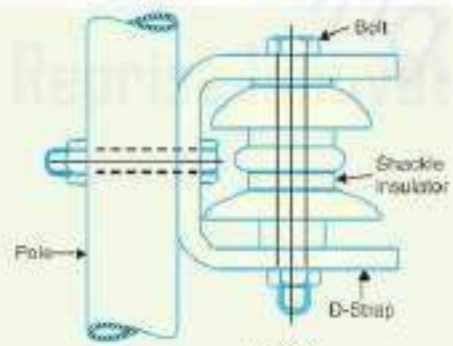


Fig. 8.9

Sag In Overhead Lines

While erecting an overhead line, it is very important that conductors are under safe tension. If the conductors are too much stretched between supports in a bid to save conductor material, the stress in the conductor may reach unsafe value and in certain cases the conductor may break due to excessive tension. In order to permit safe tension in the conductors, they are not fully stretched but are allowed to have a dip or sag.

The difference in level between points of supports and the lowest point on the conductor is called **sag.**

Fig. 8.23. (*i*) shows a conductor suspended between two equilevel supports *A* and *B*. The conductor is not fully stretched but is allowed to have a dip. The lowest point on the conductor is *O* and the sag is *S*. The following points may be noted :

Fig. 8.23

(*i*) When the conductor is suspended between two supports at the same level, it takes the shape of catenary. However, if the sag is very small compared with the span, then sag-span curve is like a parabola.

(*ii*) The tension at any point on the conductor acts tangentially. Thus tension T_O at the lowest point *O* acts horizontally as shown in Fig. 8.23. (*ii*).

(*iii*) The horizontal component of tension is constant throughout the length of the wire.

(*iv*) The tension at supports is approximately equal to the horizontal tension acting at any point on the wire. Thus if *T* is the tension at the support *B*, then $T = T_O$.

Conductor sag and tension. This is an important consideration in the mechanical design of overhead lines. The conductor sag should be kept to a minimum in order to reduce the conductor material required and to avoid extra pole height for sufficient clearance above ground level. It is also desirable that tension in the conductor should be low to avoid the mechanical failure of conductor and to permit the use of less strong supports. However, low conductor tension and minimum sag are not possible. It is because low sag means a tight wire and high tension, whereas a low tension means a loose wire and increased sag. Therefore, in actual practice, a compromise in made between the two.

Calculation of Sag

In an overhead line, the sag should be so adjusted that tension in the conductors is within safe limits. The tension is governed by conductor weight, effects of wind, ice loading and temperature variations. It is a standard practice to keep conductor tension less than 50% of its ultimate tensile strength *i.e.*, minimum factor of safety in respect of conductor tension should be 2. We shall now calculate sag and tension of a conductor when (*i*) supports are at equal levels and (*ii*) supports are at unequal levels.

(*i*) **When supports are at equal levels.** Consider a conductor between two equilevel supports A and B with O as the lowest point as shown in Fig. 8.24. It can be proved that lowest point will be at the mid-span.

Let

l = Length of span

w = Weight per unit length of conductor

T = Tension in the conductor.

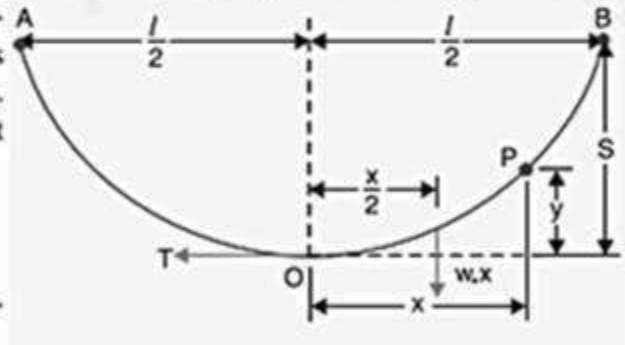


Fig. 8.24

Consider a point P on the conductor. Taking the lowest point O as the origin, let the co-ordinates of point P be x and y. Assuming that the curvature is so small that curved length is equal to its horizontal projection (*i.e.*, $OP = x$), the two forces acting on the portion OP of the conductor are :

(*a*) The weight wx of conductor acting at a distance $x/2$ from O.

(*b*) The tension T acting at O.

Equating the moments of above two forces about point O, we get,

$$T y = w x \times \frac{x}{2}$$

or $$y = \frac{w x^2}{2 T}$$

The maximum dip (sag) is represented by the value of y at either of the supports A and B.

At support A, $x = l/2$ and $y = S$

$\therefore$ $$\text{Sag,} \quad S = \frac{w(l/2)^2}{2T} = \frac{w l^2}{8 T}$$

(*ii*) **When supports are at unequal levels.** In hilly areas, we generally come across conductors suspended between supports at unequal levels. Fig. 8.25 shows a conductor suspended between two supports A and B which are at different levels. The lowest point on the conductor is O.

Let

l = Span length

h = Difference in levels between two supports

x_1 = Distance of support at lower level (*i.e.*, A) from O

x_2 = Distance of support at higher level (*i.e.* B) from O

T = Tension in the conductor

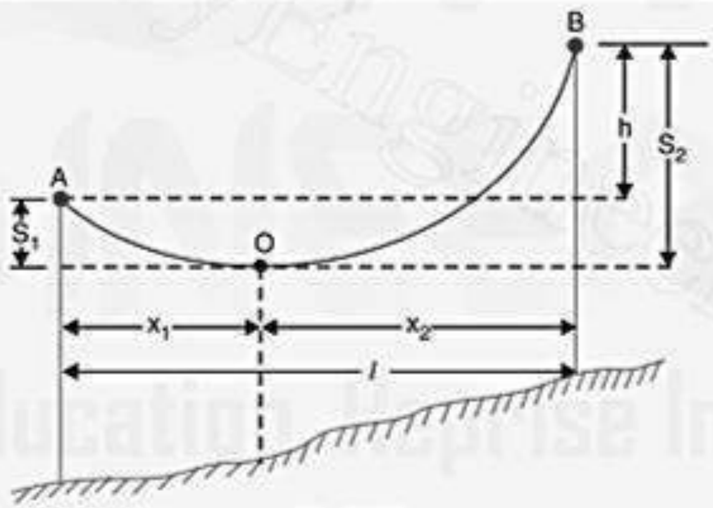


Fig. 8.25

If w is the weight per unit length of the conductor, then,

$$\text{Sag } S_1 = \frac{w x_1^{2*}}{2T}$$

and $$\text{Sag } S_2 = \frac{w x_2^2}{2T}$$

Also $$x_1 + x_2 = l \quad \text{...(}i\text{)}$$

Now $$S_2 - S_1 = \frac{w}{2T}[x_2^2 - x_1^2] = \frac{w}{2T}(x_2 + x_1)(x_2 - x_1)$$

∴ $$S_2 - S_1 = \frac{w\,l}{2T}(x_2 - x_1) \qquad [\because x_1 + x_2 = l]$$

But $$S_2 - S_1 = h$$

∴ $$h = \frac{w\,l}{2T}(x_2 - x_1)$$

or $$x_2 - x_1 = \frac{2\,Th}{w\,l} \qquad ...(ii)$$

Solving exps. (*i*) and (*ii*), we get,

$$x_1 = \frac{l}{2} - \frac{T\,h}{w\,l}$$

$$x_2 = \frac{l}{2} + \frac{T\,h}{w\,l}$$

Having found x_1 and x_2, values of S_1 and S_2 can be easily calculated.

Effect of wind and ice loading. The above formulae for sag are true only in still air and at normal temperature when the conductor is acted by its weight only. However, in actual practice, a conductor may have ice coating and simultaneously subjected to wind pressure. The weight of ice acts vertically downwards *i.e.*, in the same direction as the weight of conductor. The force due to the wind is assumed to act horizontally *i.e.*, at right angle to the projected surface of the conductor. Hence, the total force on the conductor is the vector sum of horizontal and vertical forces as shown in Fig. 8.26 (*iii*).

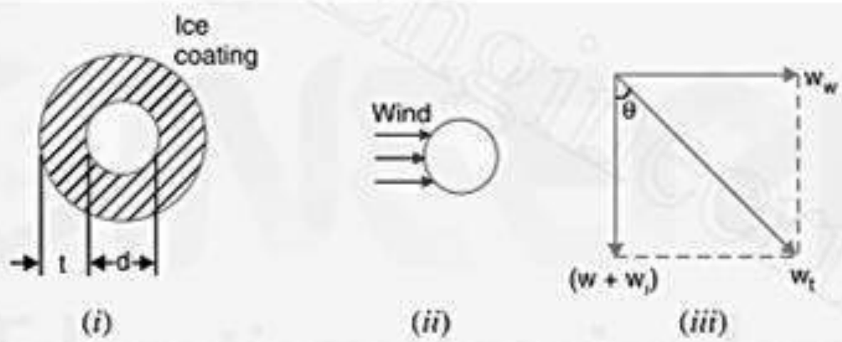


Fig. 8.26

Total weight of conductor per unit length is

$$w_t = \sqrt{(w + w_i)^2 + (w_w)^2}$$

where w = weight of conductor per unit length

= conductor material density × volume per unit length

w_i = weight of ice per unit length

= density of ice × volume of ice per unit length

$$= \text{density of ice} \times \frac{\pi}{4}\,[(d + 2t)^2 - d^2] \times 1$$

= density of ice × $\pi\, t\,(d + t)$*

w_w = wind force per unit length

= wind pressure per unit area × projected area per unit length

= wind pressure × $[(d + 2t) \times 1]$

When the conductor has wind and ice loading also, the following points may be noted :

(*i*) The conductor sets itself in a plane at an angle θ to the vertical where

$$\tan\theta = \frac{w_w}{w + w_i}$$

(*ii*) The sag in the conductor is given by :

$$S = \frac{w_t\, l^2}{2T}$$

Hence S represents the slant sag in a direction making an angle θ to the vertical. *If no specific mention is made in the problem, then slant slag is calculated by using the above formula.*

(*iii*) The vertical sag $= S\cos\theta$

Example *A 132 kV transmission line has the following data :*

Wt. of conductor = 680 kg/km ; *Length of span = 260 m*

Ultimate strength = 3100 kg ; *Safety factor = 2*

Calculate the height above ground at which the conductor should be supported. Ground clearance required is 10 metres.

Solution.

Wt. of conductor/metre run, $w = 680/1000 = 0{\cdot}68$ kg

Working tension, $T = \dfrac{\text{Ultimate strength}}{\text{Safety factor}} = \dfrac{3100}{2} = 1550$ kg

Span length, $l = 260$ m

$$\therefore \quad \text{Sag} = \frac{w\,l^2}{8T} = \frac{0\cdot 68\times(260)^2}{8\times 1550} = 3\cdot 7 \text{ m}$$

∴ Conductor should be supported at a height of $10 + 3{\cdot}7 =$ **13·7 m**

Example *A transmission line has a span of 150 m between level supports. The conductor has a cross-sectional area of 2 cm². The tension in the conductor is 2000 kg. If the specific gravity of the conductor material is 9·9 gm/cm³ and wind pressure is 1·5 kg/m length, calculate the sag. What is the vertical sag?*

Solution.

Span length, $l = 150$ m; Working tension, $T = 2000$ kg

Wind force/m length of conductor, $w_w = 1{\cdot}5$ kg

Wt. of conductor/m length, w = Sp. Gravity × Volume of 1 m conductor

$= 9{\cdot}9 \times 2 \times 100 = 1980$ gm $= 1{\cdot}98$ kg

Total weight of 1 m length of conductor is

$$w_t = \sqrt{w^2 + w_w^2} = \sqrt{(1\cdot 98)^2 + (1\cdot 5)^2} = 2\cdot 48 \text{ kg}$$

$$\therefore \quad \text{Sag, } S = \frac{w_t\, l^2}{8T} = \frac{2\cdot 48\times(150)^2}{8\times 2000} = \mathbf{3\cdot 48\ m}$$

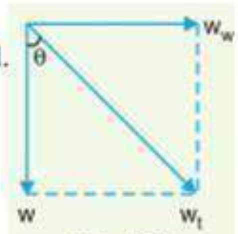


Fig. 8.27

This is the value of slant sag in a direction making an angle θ with the vertical. Referring to Fig. 8.27, the value of θ is given by ;

$$\tan\theta = w_w/w = 1\cdot 5/1\cdot 98 = 0\cdot 76$$

$$\therefore \quad \theta = \tan^{-1} 0\cdot 76 = 37\cdot 23°$$

$$\therefore \quad \text{Vertical sag} = S\cos\theta$$

$$= 3\cdot 48\times\cos 37\cdot 23° = \mathbf{2\cdot 77\ m}$$

Example *A transmission line has a span of 200 metres between level supports. The conductor has a cross-sectional area of 1·29 cm², weighs 1170 kg/km and has a breaking stress of 4218 kg/cm². Calculate the sag for a safety factor of 5, allowing a wind pressure of 122 kg per square metre of projected area. What is the vertical sag?*

Solution.

Span length, $l = 200$ m

Wt. of conductor/m length, $w = 1170/1000 = 1{\cdot}17$ kg

Working tension, $^*T = 4218 \times 1{\cdot}29/5 = 1088$ kg

Diameter of conductor, $d = \sqrt{\dfrac{4 \times \text{area}}{\pi}} = \sqrt{\dfrac{4 \times 1{\cdot}29}{\pi}} = 1{\cdot}28$ cm

Wind force/m length, w_w = Pressure × projected area in m²

$= (122) \times (1{\cdot}28 \times 10^{-2} \times 1) = 1{\cdot}56$ kg

Total weight of conductor per metre length is

$$w_t = \sqrt{w^2 + w_w^2} = \sqrt{(1{\cdot}17)^2 + (1{\cdot}56)^2} = 1{\cdot}95 \text{ kg}$$

$\therefore$ Slant sag, $S = \dfrac{w_t\, l^2}{8T} = \dfrac{1{\cdot}95 \times (200)^2}{8 \times 1088} = \mathbf{8{\cdot}96\ m}$

The slant sag makes an angle θ with the vertical where value of θ is given by :

$$\theta = \tan^{-1}(w_w/w) = \tan^{-1}(1{\cdot}56/1{\cdot}17) = 53{\cdot}13°$$

$\therefore$ Vertical sag $= S \cos\theta = 8{\cdot}96 \times \cos 53{\cdot}13° = \mathbf{5{\cdot}37\ m}$

Example *A transmission line has a span of 275 m between level supports. The conductor has an effective diameter of 1·96 cm and weighs 0·865 kg/m. Its ultimate strength is 8060 kg. If the conductor has ice coating of radial thickness 1·27 cm and is subjected to a wind pressure of 3·9 gm/cm² of projected area, calculate sag for a safety factor of 2. Weight of 1 c.c. of ice is 0·91 gm.*

Solution.

Span length, $l = 275$ m ; Wt. of conductor/m length, $w = 0{\cdot}865$ kg

Conductor diameter, $d = 1{\cdot}96$ cm ; Ice coating thickness, $t = 1{\cdot}27$ cm

Working tension, $T = 8060/2 = 4030$ kg

Volume of ice per metre (*i.e.*, 100 cm) length of conductor

$= \pi\, t\,(d + t) \times 100$ cm³

$= \pi \times 1{\cdot}27 \times (1{\cdot}96 + 1{\cdot}27) \times 100 = 1288$ cm³

Weight of ice per metre length of conductor is

$$w_i = 0{\cdot}91 \times 1288 = 1172 \text{ gm} = 1{\cdot}172 \text{ kg}$$

Wind force/m length of conductor is

$w_w = [\text{Pressure}] \times [(d + 2t) \times 100]$

$= [3{\cdot}9] \times (1{\cdot}96 + 2 \times 1{\cdot}27) \times 100$ gm $= 1755$ gm $= 1{\cdot}755$ kg

Total weight of conductor per metre length of conductor is

$$w_t = \sqrt{(w + w_i)^2 + (w_w)^2}$$

$$= \sqrt{(0{\cdot}865 + 1{\cdot}172)^2 + (1{\cdot}755)^2} = 2{\cdot}688 \text{ kg}$$

$\therefore$ $$\text{Sag} = \frac{w_t l^2}{8T} = \frac{2\cdot688 \times (275)^2}{8 \times 4030} = \mathbf{6\cdot3\ m}$$

Example *A transmission line has a span of 214 metres between level supports. The conductors have a cross-sectional area of 3·225 cm². Calculate the factor of safety under the following conditions :*

Vertical sag = 2·35 m ; *Wind pressure = 1·5 kg/m run*

Breaking stress = 2540 kg/cm² ; *Wt. of conductor = 1·125 kg/m run*

Solution.

Here, $l = 214\text{ m}$; $w = 1\cdot125\text{ kg}$; $w_w = 1\cdot5\text{ kg}$

Total weight of one metre length of conductor is

$$w_t = \sqrt{w^2 + w_w^2} = \sqrt{(1\cdot125)^2 + (1\cdot5)^2} = 1\cdot875\text{ kg}$$

If f is the factor of safety, then,

Working tension, $$T = \frac{\text{Breaking stress} \times \text{conductor area}}{\text{safety factor}} = 2540 \times 3\cdot225/f = 8191/f\text{ kg}$$

$$\text{Slant Sag,}\quad S = \frac{\text{Vertical sag}}{{}^{*}\cos\theta} = \frac{2\cdot35 \times 1\cdot875}{1\cdot125} = 3\cdot92\text{ m}$$

Now $$S = \frac{w_t l^2}{8T}$$

or $$T = \frac{w_t l^2}{8S}$$

$\therefore$ $$\frac{8191}{f} = \frac{1\cdot875 \times (214)^2}{8 \times 3\cdot92}$$

or Safety factor, $$f = \frac{8191 \times 8 \times 3\cdot92}{1\cdot875 \times (214)^2} = \mathbf{3}$$

Example *An overhead line has a span of 150 m between level supports. The conductor has a cross-sectional area of 2 cm². The ultimate strength is 5000 kg/cm² and safety factor is 5. The specific gravity of the material is 8·9 gm/cc. The wind pressure is 1·5 kg/m. Calculate the height of the conductor above the ground level at which it should be supported if a minimum clearance of 7 m is to be left between the ground and the conductor.*

Solution.

Span length, $l = 150$ m ; Wind force/m run, $w_w = 1\cdot5$ kg

Wt. of conductor/m run, w = conductor area × 100 cm × sp. gravity

$= 2 \times 100 \times 8\cdot9 = 1780\text{ gm} = 1\cdot78\text{ kg}$

Working tension, $T = 5000 \times 2/5 = 2000\text{ kg}$

Total weight of one metre length of conductor is

$$w_t = \sqrt{w^2 + w_w^2} = \sqrt{(1\cdot78)^2 + (1\cdot5)^2} = 2\cdot33\text{ kg}$$

$$\text{Slant sag,}\quad S = \frac{w_t l^2}{8T} = \frac{2\cdot33 \times (150)^2}{8 \times 2000} = 3\cdot28\text{ m}$$

$$\text{Vertical sag} = S\cos\theta = 3\cdot28 \times w/w_t = 3\cdot28 \times 1\cdot78/2\cdot33 = 2\cdot5\text{ m}$$

Conductor should be supported at a height of $7 + 2\cdot5 = \mathbf{9\cdot5\ m}$

Example *The towers of height 30 m and 90 m respectively support a transmission line conductor at water crossing. The horizontal distance betwen the towers is 500 m. If the tension in the conductor is 1600 kg, find the minimum clearance of the conductor and water and clearance mid-way between the supports. Weight of conductor is 1·5 kg/m. Bases of the towers can be considered to be at water level.*

Solution. Fig. 8.28 shows the conductor suspended between two supports A and B at different levels with O as the lowest point on the conductor.

Here, $l = 500$ m ; $w = 1{\cdot}5$ kg ; $T = 1600$ kg.

Difference in levels between supports, $h = 90 - 30 = 60$ m. Let the lowest point O of the conductor be at a distance x_1 from the support at lower level (*i.e.*, support A) and at a distance x_2 from the support at higher level (*i.e.*, support B).

Obviously, $\quad x_1 + x_2 = 500$ m $\quad$ ...(*i*)

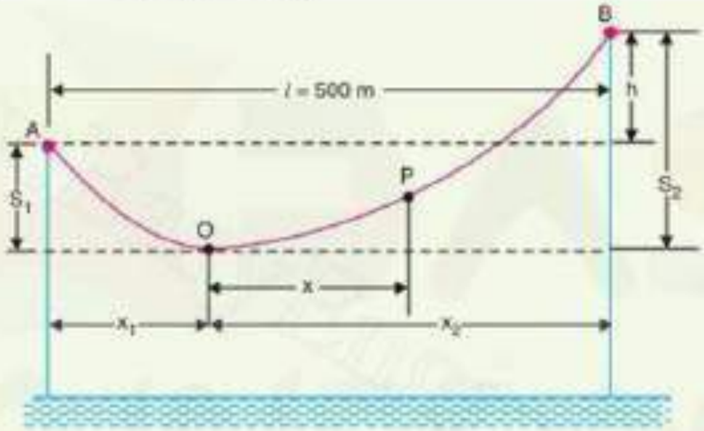


Fig. 8.28

Now $\quad$ Sag $S_1 = \dfrac{w\,x_1^2}{2T}$ and Sag $S_2 = \dfrac{w\,x_2^2}{2T}$

$$\therefore \quad h = S_2 - S_1 = \frac{w\,x_2^2}{2T} - \frac{w\,x_1^2}{2T}$$

or

$$60 = \frac{w}{2T}(x_2 + x_1)(x_2 - x_1)$$

$$\therefore \quad x_2 - x_1 = \frac{60 \times 2 \times 1600}{1{\cdot}5 \times 500} = 256 \text{ m} \quad \text{...(ii)}$$

Solving exps. (*i*) and (*ii*), we get, $x_1 = 122$ m; $x_2 = 378$ m

Now, $\quad S_1 = \dfrac{w\,x_1^2}{2T} = \dfrac{1{\cdot}5 \times (122)^2}{2 \times 1600} = 7$ m

Clearance of the lowest point O from water level

$= 30 - 7 =$ **23 m**

Let the mid-point P be at a distance x from the lowest point O.

Clearly, $\quad x = 250 - x_1 = 250 - 122 = 128$ m

Sag at mid-point P, $\quad S_{mid} = \dfrac{w\,x^2}{2T} = \dfrac{1{\cdot}5 \times (128)^2}{2 \times 1600} = 7{\cdot}68$ m

Clearance of mid-point P from water level

$$= 23 + 7{\cdot}68 = \mathbf{30{\cdot}68\ m}$$

Example *An overhead transmission line conductor having a parabolic configuration weighs 1·925 kg per metre of length. The area of X-section of the conductor is 2·2 cm² and the ultimate strength is 8000 kg/cm². The supports are 600 m apart having 15 m difference of levels. Calculate the sag from the taller of the two supports which must be allowed so that the factor of safety shall be 5. Assume that ice load is 1 kg per metre run and there is no wind pressure.*

Solution. Fig. 8.29. shows the conductor suspended between two supports at A and B at different levels with O as the lowest point on the conductor.

Here,
$$l = 600\ \text{m} ;\ w_i = 1\ \text{kg} ;\ h = 15\text{m}$$
$$w = 1{\cdot}925\ \text{kg} ;\ T = 8000 \times 2{\cdot}2/5 = 3520\ \text{kg}$$

Total weight of 1 m length of conductor is

$$w_t = w + w_i = 1{\cdot}925 + 1 = 2{\cdot}925\ \text{kg}$$

Let the lowest point O of the conductor be at a distance x_1 from the support at lower level (*i.e.*, A) and at a distance x_2 from the support at higher level (*i.e.*, B).

Clearly, $$x_1 + x_2 = 600\ \text{m} \qquad ...(i)$$

Now, $$h = S_2 - S_1 = \frac{w_t x_2^2}{2T} - \frac{w_t x_1^2}{2T}$$

or $$15 = \frac{w_t}{2T}(x_2 + x_1)(x_2 - x_1)$$

$\therefore$ $$x_2 - x_1 = \frac{2 \times 15 \times 3520}{2{\cdot}925 \times 600} = 60\ \text{m} \qquad ...(ii)$$

Solving exps. (*i*) and (*ii*), we have, $x_1 = 270$ m and $x_2 = 330$ m

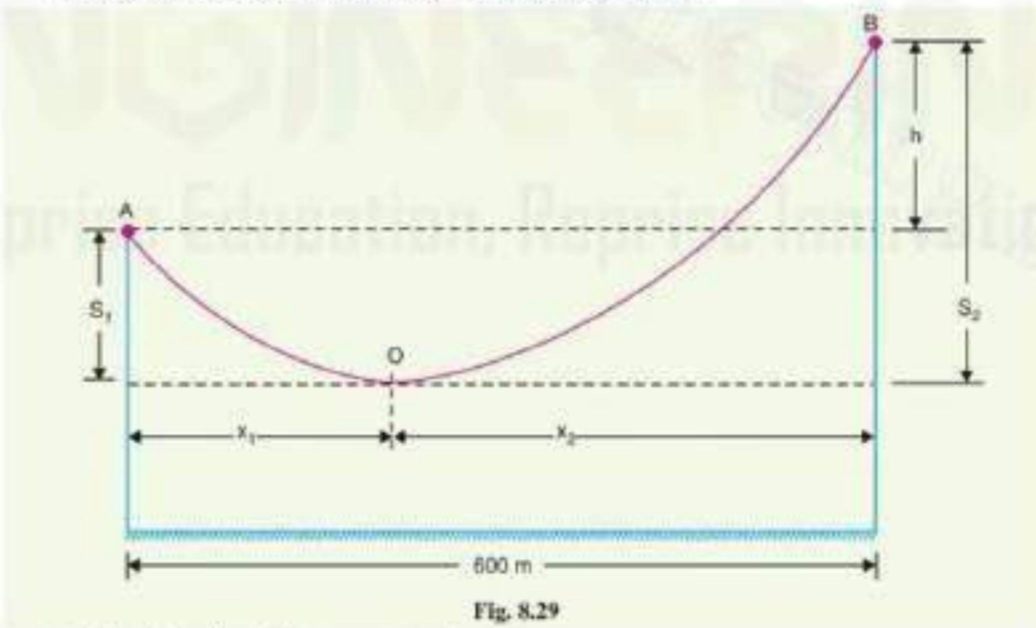


Fig. 8.29

Sag from the taller of the two towers is

$$S_2 = \frac{w_t x_2^2}{2T} = \frac{2{\cdot}925 \times (330)^2}{2 \times 3520} = \mathbf{45{\cdot}24\ m}$$

Example *An overhead transmission line at a river crossing is supported from two towers at heights of 40 m and 90 m above water level, the horizontal distance between the towers being 400 m. If the maximum allowable tension is 2000 kg, find the clearance between the conductor and water at a point mid-way between the towers. Weight of conductor is 1 kg/m.*

Solution. Fig. 8.30 shows the whole arrangement.

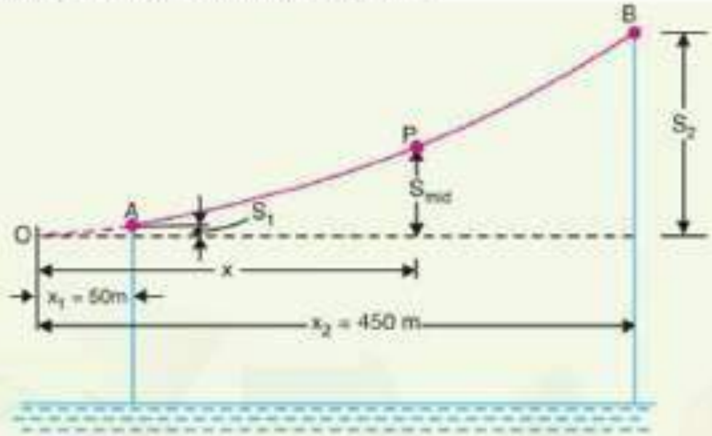


Fig. 8.30

Here,
$$h = 90 - 40 = 50 \text{ m}; \quad l = 400 \text{ m}$$
$$T = 2000 \text{ kg}; \quad w = 1 \text{ kg/m}$$

Obviously,
$$x_1 + x_2 = 400 \text{ m} \quad \text{...(i)}$$

Now
$$h = S_2 - S_1 = \frac{wx_2^2}{2T} - \frac{wx_1^2}{2T}$$

or
$$50 = \frac{w}{2T}(x_2 + x_1)(x_2 - x_1)$$

$\therefore$
$$x_2 - x_1 = \frac{50 \times 2 \times 2000}{400} = 500 \text{ m} \quad \text{...(ii)}$$

Solving exps. (*i*) and (*ii*), we get, $x_2 = 450$ m and $x_1 = -50$ m

Now x_2 is the distance of higher support B from the lowest point O on the conductor, whereas x_1 is that of lower support A. As the span is 400 m, therefore, point A lies on the same side of O as B (see Fig. 8.30).

Horizontal distance of mid-point P from lowest point O is
$$x = \text{Distance of } A \text{ from } O + 400/2 = 50 + 200 = 250 \text{ m}$$

$\therefore$ Sag at point P,
$$S_{mid} = \frac{w\,x^2}{2T} = \frac{1 \times (250)^2}{2 \times 2000} = 15{\cdot}6 \text{ m}$$

Now
$$\text{Sag } S_2 = \frac{w\,x_2^2}{2T} = \frac{1 \times (450)^2}{2 \times 2000} = 50{\cdot}6 \text{ m}$$

Height of point B above mid-point P
$$= S_2 - S_{mid} = 50{\cdot}6 - 15{\cdot}6 = 35 \text{ m}$$

$\therefore$ Clearance of mid-point P above water level
$$= 90 - 35 = \mathbf{55\ m}$$

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Prepared By:-

Chapter 4 performance of short and medium lines

Classification of Overhead Transmission Lines:

A transmission line has *three constants *R*, *L* and *C* distributed uniformly along the whole length of the line. The resistance and inductance form the series impedance. The capacitance existing between conductors for 1-phase line or from a conductor to neutral for a 3-phase line forms a shunt path throughout thelength oftheline. Therefore, capacitance effects introduce complications in transmission line calculations. Depending upon the manner in which capacitance is taken into account, the overhead transmission lines are classified as:

Short transmission lines. When the length of an overhead transmission line is upto about 50 km and the line voltage is comparatively low (< 20 kV), it is usually considered as a short transmission line. Due to smaller length and lower voltage, the capacitance effects are small and hence can be neglected. Therefore, while studying the performance of a short transmission line, only resistance and inductance of the line are taken into account.

Medium transmission lines. When the length of an overhead transmission line is about 50- 150 km and the line voltage is moderately high (>20 kV < 100 kV), it is considered as a medium transmission line. Due to sufficient length and voltage of the line, the capacitance effects are taken into account. For purposes of calculations, the distributed capacitance of the line is divided and lumped in the form of condensers shunted across the line at one or more points.

Long transmission lines. When the length of an overhead transmission line is more than 150 km and line voltage is very high (> 100 kV), it is considered as a long transmission line. For the treatment of such a line, the line constants are considered uniformly distributed over the whole length of the line and rigorous methods are employed for solution.

Important Terms:

While studying the performance of a transmission line, it is desirable to determine its voltage regula- tion and transmission efficiency. We shall explain these two terms in turn.

1. **Voltage regulation.** When a transmission line is carrying current, there is a voltage drop in the line due to resistance and inductance of the line. The result is that receiving end voltage (V_R) of the line is generally less than the sending end voltage (V_S). This voltage drop ($V_S - V_R$) in the line is expressed as a percentage of receiving end voltage V_R and is called voltage regulation.

*The difference in voltage at the receiving end of a transmission line **between conditions of no load and full load is called* **voltage regulation** and is expressed as a percentage of the receiving end voltage.

Mathematically %age of voltage regulation = V_s -V_R / V_R X100

Obviously, it is desirable that the voltage regulation of a transmission line should be low *i.e.*, the increase in load current should make very little difference in the receiving end voltage.

2. **Transmission efficiency.** The power obtained at the receiving end of a transmission line is generally less than the sending end power due to losses in the line resistance.

The ratio of receiving end power to the sending end power of a transmission line is known as the

transmission efficiency *of the line i.e.* % age Transmission efficiency, η_T

$$= \frac{\text{Receiving end power}}{\text{Sending end power}} \times 100$$

$= \left(V_R I_R \cos \phi_R \;/\; V_S I_S \cos \phi_S \right) \times 100$

where V_R, I_R and $\cos \phi_R$ are the receiving end voltage, current and power factor while V_S, I_S and $\cos \phi_S$ are the corresponding values at the sending end.

Performance of Single Phase Short Transmission Lines

As stated earlier, the effects of line capacitance are neglected for a short transmission line. Therefore, while studying the performance of such a line, only resistance and inductance of the line are taken into account. The equivalent circuit of a single phase short transmission line is shown in Fig. 10.1 (*i*). Here, the total line resistance and inductance are shown as concentrated or lumped instead of being distributed. The circuit is a simple a.c. series circuit.

Let

I = load current

R = loop resistance *i.e.*, resistance of both conductors

X_L = loop reactance

V_R = receiving end voltage

$\cos \phi_R$ = receiving end power factor (lagging)

V_S = sending end voltage

$\cos \phi_S$ = sending end power factor

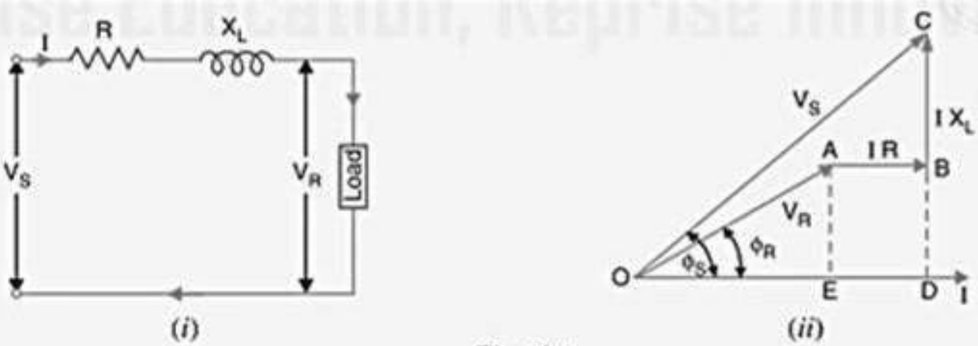


Fig. 10.1

The *phasor diagram of the line for lagging load power factor is shown in Fig. 10.1 (*ii*). From the right angled traingle *ODC*, we get,

$$(OC)^2 = (OD)^2 + (DC)^2$$

or
$$V_S^2 = (OE + ED)^2 + (DB + BC)^2$$
$$= (V_R \cos \phi_R + IR)^2 + (V_R \sin \phi_R + IX_L)^2$$

$\therefore$
$$V_S = \sqrt{(V_R \cos \phi_R + IR)^2 + (V_R \sin \phi_R + IX_L)^2}$$

(*i*) %age Voltage regulation $= \dfrac{V_S - V_R}{V_R} \times 100$

(*ii*) Sending end *p.f.*, $\cos \phi_S = \dfrac{OD}{OC} = \dfrac{V_R \cos \phi_R + IR}{V_S}$

(*iii*) Power delivered $= V_R I_R \cos \phi_R$

Line losses $= I^2 R$

Power sent out $= V_R I_R \cos \phi_R + I^2 R$

%age Transmission efficiency $= \dfrac{\text{Power delivered}}{\text{Power sent out}} \times 100$

$$= \frac{V_R\, I_R \cos \phi_R}{V_R\, I_R \cos \phi_R + I^2 R} \times 100$$

An approximate expression for the sending end voltage V_S can be obtained as follows. Draw perpendicular from *B* and *C* on *OA* produced as shown in Fig. 10.2. Then *OC* is *nearly* equal to *OF i.e.*,

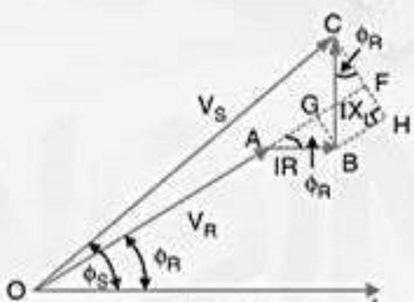


Fig. 10.2

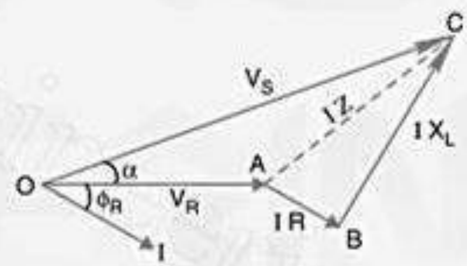


Fig. 10.3

$$OC = OF = OA + AF = OA + AG + GF$$
$$= OA + AG + BH$$

$\therefore$
$$V_S = V_R + I\,R \cos \phi_R + I\,X_L \sin \phi_R$$

Solution in complex notation. It is often convenient and profitable to make the line calculations in complex notation.

Taking $\overrightarrow{V_R}$ as the reference phasor, draw the phasor diagram as shown in Fig 10.3. It is clear that $\overrightarrow{V_S}$ is the phasor sum of $\overrightarrow{V_R}$ and $\vec{I}\,\vec{Z}$.

$$^*\overrightarrow{V_R} = V_R + j\,0$$
$$\vec{I} = \vec{I}\ \angle -\phi_R = I\,(\cos \phi_R - j \sin \phi_R)$$
$$\vec{Z} = R + jX_L$$

$\therefore$
$$\overrightarrow{V_S} = \overrightarrow{V_R} + \vec{I}\,\vec{Z}$$
$$= (V_R + j\,0) + I\,(\cos \phi_R - j \sin \phi_R)\,(R + j\,X_L)$$

$$= (V_R + I R \cos \phi_R + I X_L \sin \phi_R) + j\,(I X_L \cos \phi_R - I R \sin \phi_R)$$

$$\therefore \qquad V_S = \sqrt{(V_R + I R \cos \phi_R + I X_L \sin \phi_R)^2 + (I X_L \cos \phi_R - I R \sin \phi_R)^2}$$

The second term under the root is quite small and can be neglected with reasonable accuracy. Therefore, approximate expression for V_S becomes :

$$V_S = V_R + I R \cos \phi_R + I X_L \sin \phi_R$$

The following poins may be noted :

(*i*) The approximate formula for V_S $(= V_R + I R \cos \phi_R + I X_L \sin \phi_R)$ gives fairly correct results for lagging power factors. However, appreciable error is caused for leading power factors. Therefore, approximate expression for V_S should be used for lagging p.f. only.

(*ii*) The solution in complex notation is in more presentable form.

Three-Phase Short Transmission Lines

For reasons associated with economy, transmission of electric power is done by 3-phase system. This system may be regarded as consisting of three single phase units, each wire transmitting one-third of the total power. As a matter of convenience, we generally analyse 3-phase system by considering

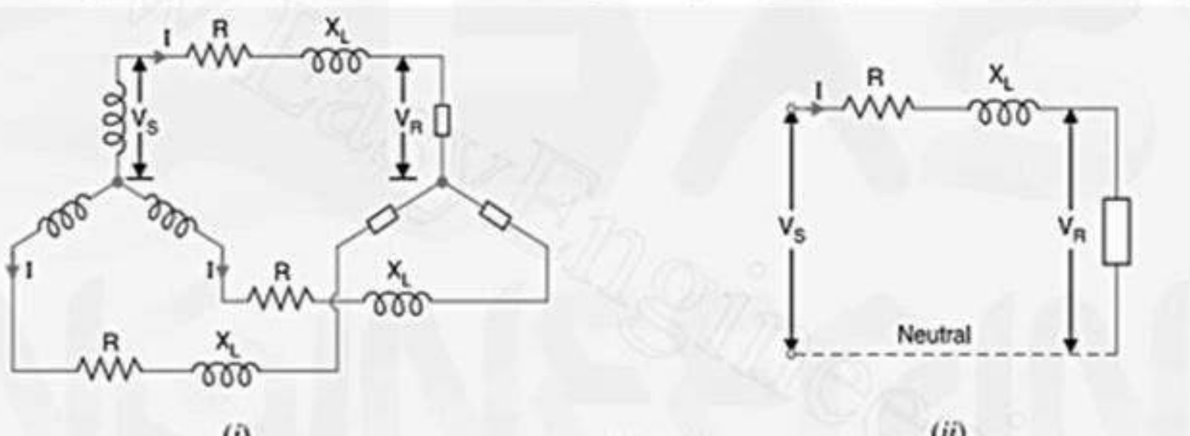


Fig. 10.4

*one phase only. Therefore, expression for regulation, efficiency etc. derived for a single phase line can also be applied to a 3-phase system. Since only one phase is considered, phase values of 3-phase system should be taken. Thus, V_S and V_R are the phase voltages, whereas R and X_L are the resistance and inductive reactance per phase respectively.

Fig. 10.4 (*i*) shows a *Y*-connected generator supplying a balanced *Y*-connected load through a transmission line. Each conductor has a resistance of $R\,\Omega$ and inductive reactance of $X_L\,\Omega$. Fig. 10.4 (*ii*) shows one phase separately. The calculations can now be made in the same way as for a single phase line.

Effect of Load p.f. on Regulation and Efficiency

The regulation and efficiency of a transmission line depend to a considerable extent upon the power factor of the load.

1. Effect on regulation. The expression for voltage regulation of a short transmission line is given by :

$$\%\text{age Voltage regulation} = \frac{I R \cos \phi_R + I X_L \sin \phi_R}{V_R} \times 100 \qquad \textit{(for lagging p.f.)}$$

$$\%\text{age Voltage regulation} = \frac{I R \cos \phi_R - I X_L \sin \phi_R}{V_R} \times 100 \qquad \textit{(for leading p.f.)}$$

The following conclusions can be drawn from the above expressions :

(*i*) When the load p.f. is lagging or unity or such leading that $I R \cos \phi_R > I X_L \sin \phi_R$, then voltage regulation is positive *i.e.*, receiving end voltage V_R will be less than the sending end voltage V_S.

(*ii*) For a given V_R and I, the voltage regulation of the line increases with the decrease in p.f. for lagging loads.

(*iii*) When the load p.f. is leading to this extent that $I X_L \sin \phi_R > I R \cos \phi_R$, then voltage regulation is negative *i.e.* the receiving end voltage V_R is more than the sending end voltage V_S.

(*iv*) For a given V_R and I, the voltage regulation of the line decreases with the decrease in p.f. for leading loads.

2. Effect on transmission efficiency. The power delivered to the load depends upon the power factor.

$$P = V_R * I \cos \phi_R \quad \text{(For 1-phase line)}$$

$$\therefore \quad I = \frac{P}{V_R \cos \phi_R}$$

$$P = 3\, V_R\, I \cos \phi_R \quad \text{(For 3-phase line)}$$

$$\therefore \quad I = \frac{P}{3 V_R \cos \phi_R}$$

Power Factor Meter

It is clear that in each case, for a given amount of power to be transmitted (P) and receiving end voltage (V_R), the load current I is inversely proportional to the load p.f. $\cos \phi_R$. Consequently, with the decrease in load p.f., the load current and hence the line losses are increased. This leads to the conclusion that transmission efficiency of a line decreases with the decrease in load p.f. and *vice-versa*.


Power Factor Regulator

Example *A single phase overhead transmission line delivers 1100 kW at 33 kV at 0·8 p.f. lagging. The total resistance and inductive reactance of the line are 10 Ω and 15 Ω respectively. Determine :* (*i*) *sending end voltage* (*ii*) *sending end power factor and* (*iii*) *transmission efficiency.*

Solution.

Load power factor, $\cos \phi_R = 0{\cdot}8$ lagging

Total line impedance, $\vec{Z} = R + j X_L = 10 + j\, 15$

Receiving end voltage, $V_R = 33\text{ kV} = 33{,}000\text{ V}$

$$\therefore \quad \text{Line current, } I = \frac{kW \times 10^3}{V_R \cos\phi_R} = \frac{1100 \times 10^3}{33{,}000 \times 0{\cdot}8} = 41{\cdot}67\text{ A}$$

$$\text{As } \cos\phi_R = 0{\cdot}8 \quad \therefore \quad \sin\phi_R = 0{\cdot}6$$

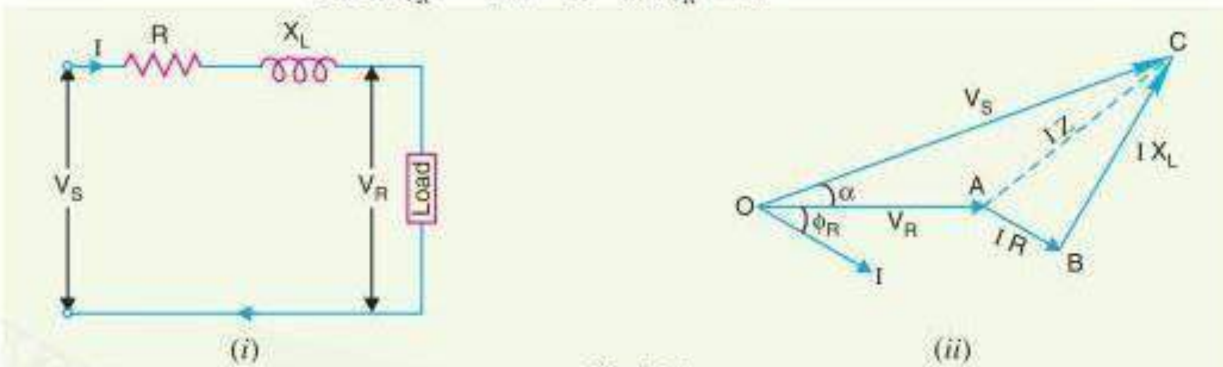


Fig. 10.5

The equivalent circuit and phasor diagram of the line are shown in Figs. 10.5 (*i*) and 10.5 (*ii*) respectively. Taking receiving end voltage $\overline{V_R}$ as the reference phasor,

$$\overline{V_R} = V_R + j\,0 = 33000\text{ V}$$

$$\vec{I} = I(\cos\phi_R - j\sin\phi_R)$$

$$= 41{\cdot}67\,(0{\cdot}8 - j\,0{\cdot}6) = 33{\cdot}33 - j\,25$$

(*i*) Sending end voltage, $\overline{V_S} = \overrightarrow{V_R} + \vec{I}\,\vec{Z}$

$$= 33{,}000 + (33{\cdot}33 - j\,25{\cdot}0)\,(10 + j\,15)$$

$$= 33{,}000 + 333{\cdot}3 - j250 + j500 + 375$$

$$= 33{,}708{\cdot}3 + j\,250$$

$$\therefore \quad \text{Magnitude of } V_S = \sqrt{(33{,}708{\cdot}3)^2 + (250)^2} = \mathbf{33{,}709\text{ V}}$$

(*ii*) Angle between $\overline{V_S}$ and $\overline{V_R}$ is

$$\alpha = \tan^{-1}\frac{250}{33{,}708{\cdot}3} = \tan^{-1}0{\cdot}0074 = 0{\cdot}42°$$

$\therefore$ Sending end power factor angle is

$$\phi_S = \phi_R + \alpha = 36{\cdot}87° + 0{\cdot}42° = 37{\cdot}29°$$

$\therefore$ Sending end p.f., $\cos\phi_S = \cos 37{\cdot}29° = \mathbf{0{\cdot}7956\ lagging}$

(*iii*) Line losses $= I^2R = (41{\cdot}67)^2 \times 10 = 17{,}364\text{ W} = 17{\cdot}364\text{ kW}$

Output delivered $= 1100\text{ kW}$

Power sent $= 1100 + 17{\cdot}364 = 1117{\cdot}364\text{ kW}$

$$\therefore \quad \text{Transmission efficiency} = \frac{\text{Power delivered}}{\text{Power sent}} \times 100 = \frac{1100}{1117{\cdot}364} \times 100 = \mathbf{98{\cdot}44\%}$$

Note. V_S and ϕ_S can also be calculated as follows :

$$V_S = V_R + I\,R\cos\phi_R + I\,X_L\sin\phi_R \text{ (approximately)}$$

$$= 33{,}000 + 41{\cdot}67 \times 10 \times 0{\cdot}8 + 41{\cdot}67 \times 15 \times 0{\cdot}6$$

$$= 33{,}000 + 333{\cdot}36 + 375{\cdot}03$$

$$= 33708{\cdot}39\text{ V which is approximately the same as above}$$

$$\cos\phi_S = \frac{V_R\cos\phi_R + I\,R}{V_S} = \frac{33{,}000 \times 0{\cdot}8 + 41{\cdot}67 \times 10}{33{,}708{\cdot}39} = \frac{26{,}816{\cdot}7}{33{,}708{\cdot}39}$$

$$= 0{\cdot}7958$$

As stated earlier, this method gives fairly correct results for lagging p.f. The reader will find that this method is used in the solution of some numericals.

Example *What is the maximum length in km for a 1-phase transmission line having copper conductor of 0·775 cm² cross-section over which 200 kW at unity power factor and at 3300V are to be delivered ? The efficiecny of transmission is 90%. Take specific resistance as 1.725 μ Ω cm.*

Solution.

$$\text{Receiving end power} = 200 \text{ kW} = 2{,}00{,}000 \text{ W}$$

$$\text{Transmission efficiency} = 0{\cdot}9$$

$$\therefore \quad \text{Sending end power} = \frac{2{,}00{,}000}{0 \cdot 9} = 2{,}22{,}222 \text{ W}$$

$$\therefore \quad \text{Line losses} = 2{,}22{,}222 - 2{,}00{,}000 = 22{,}222 \text{ W}$$

$$\text{Line current, } I = \frac{200 \times 10^3}{3{,}300 \times 1} = 60{\cdot}6 \text{ A}$$

Let R Ω be the resistance of one conductor.

$$\text{Line losses} = 2 I^2 R$$

or

$$22{,}222 = 2\,(60{\cdot}6)^2 \times R$$

$$\therefore \quad R = \frac{22{,}222}{2 \times (60 \cdot 6)^2} = 3{\cdot}025\ \Omega$$

Now,

$$R = \rho\, l/a$$

$$\therefore \quad l = \frac{Ra}{\rho} = \frac{3 \cdot 025 \times 0 \cdot 775}{1 \cdot 725 \times 10^{-6}} = 1{\cdot}36 \times 10^6 \text{ cm} = \mathbf{13{\cdot}6 \text{ km}}$$

Example *An overhead 3-phase transmission line delivers 5000 kW at 22 kV at 0·8 p.f. lagging. The resistance and reactance of each conductor is 4 Ω and 6 Ω respectively. Determine : (i) sending end voltage (ii) percentage regulation (iii) transmission efficiency.*

Solution.

$$\text{Load power factor, } \cos\phi_R = 0{\cdot}8 \text{ lagging}$$

$$\text{Receiving end voltage/phase, }^* V_R = 22{,}000/\sqrt{3} = 12{,}700 \text{ V}$$

$$\text{Impedance/phase,} \quad \vec{Z} = 4 + j\,6$$

$$\text{Line current,} \quad I = \frac{5000 \times 10^3}{3 \times 12700 \times 0 \cdot 8} = 164 \text{ A}$$

$$\text{As } \cos\phi_R = 0{\cdot}8 \quad \therefore \sin\phi_R = 0{\cdot}6$$

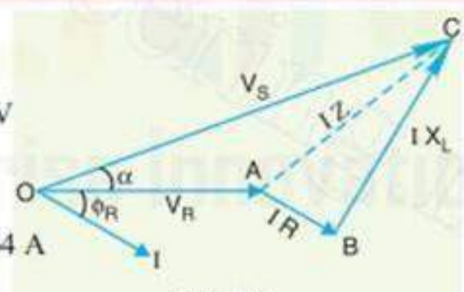


Fig. 10.6

Taking $\overline{V_R}$ as the reference phasor (see Fig. 10.6),

$$\overline{V_R} = V_R + j\,0 = 12700 \text{ V}$$

$$\vec{I} = I\,(\cos\phi_R - j\sin\phi_R) = 164\,(0{\cdot}8 - j\,0{\cdot}6) = 131{\cdot}2 - j\,98{\cdot}4$$

(*i*) Sending end voltage per phase is

$$\overline{V_S} = \overline{V_R} + \vec{I}\,\vec{Z} = 12700 + (131{\cdot}2 - j\,98{\cdot}4)\,(4 + j\,6)$$

$$= 12700 + 524{\cdot}8 + j\,787{\cdot}2 - j\,393{\cdot}6 + 590{\cdot}4$$

$$= 13815{\cdot}2 + j\,393{\cdot}6$$

$$\text{Magnitude of } V_S = \sqrt{(13815 \cdot 2)^2 + (393 \cdot 6)^2} = 13820{\cdot}8 \text{ V}$$

Line value of $V_S = \sqrt{3} \times 13820 \cdot 8 = 23938$ V = **23·938 kV**

(ii) $$\%\text{age Regulation} = \frac{V_S - V_R}{V_R} \times 100 = \frac{13820 \cdot 8 - 12700}{12700} \times 100 = \mathbf{8 \cdot 825\%}$$

(iii) $$\text{Line losses} = 3I^2R = 3 \times (164)^2 \times 4 = 3{,}22{,}752 \text{ W} = 322 \cdot 752 \text{ kW}$$

$$\therefore \quad \text{Transmission efficiency} = \frac{5000}{5000 + 322 \cdot 752} \times 100 = \mathbf{93 \cdot 94\%}$$

Example *Estimate the distance over which a load of 15000 kW at a p.f. 0·8 lagging can be delivered by a 3-phase transmission line having conductors each of resistance 1 Ω per kilometre. The voltage at the receiving end is to be 132 kV and the loss in the transmission is to be 5%.*

Solution.

Line current, $$I = \frac{\text{Power delivered}}{\sqrt{3} \times \text{line voltage} \times \text{power factor}} = \frac{15000 \times 10^3}{\sqrt{3} \times 132 \times 10^3 \times 0 \cdot 8} = 82 \text{ A}$$

$$\text{Line losses} = 5\% \text{ of power delivered} = 0 \cdot 05 \times 15000 = 750 \text{ kW}$$

Let R Ω be the resistance of one conductor.

$$\text{Line losses} = 3\,I^2R$$

or $$750 \times 10^3 = 3 \times (82)^2 \times R$$

$$\therefore \quad R = \frac{750 \times 10^3}{3 \times (82)^2} = 37 \cdot 18\,\Omega$$

Resistance of each conductor per km is 1 Ω (*given*).

∴ Length of line = **37·18 km**

Example *A 3-phase line delivers 3600 kW at a p.f. 0·8 lagging to a load. If the sending end voltage is 33 kV, determine (i) the receiving end voltage (ii) line current (iii) transmission efficiency. The resistance and reactance of each conductor are 5·31 Ω and 5·54 Ω respectively.*

Solution.

Resistance of each conductor, $R = 5 \cdot 31\,\Omega$

Reactance of each conductor, $X_L = 5 \cdot 54\,\Omega$

Load power factor, $\cos\phi_R = 0 \cdot 8$ (*lagging*)

Sending end voltage/phase, $V_S = 33{,}000/\sqrt{3} = 19{,}052$ V

Let V_R be the phase voltage at the receiving end.

Line current, $$I = \frac{\text{Power delivered / phase}}{V_R \times \cos\phi_R} = \frac{1200 \times 10^3}{V_R \times 0 \cdot 8}$$

$$= \frac{150 \times 10^5}{V_R} \qquad \text{...}(i)$$

(*i*) Using approximate expression for V_S, we get,

$$V_S = V_R + I\,R\cos\phi_R + I\,X_L \sin\phi_R$$

or $$19{,}052 = V_R + \frac{15 \times 10^5}{V_R} \times 5 \cdot 31 \times 0 \cdot 8 + \frac{15 \times 10^5}{V_R} \times 5 \cdot 54 \times 0 \cdot 6$$

or $$V_R^2 - 19{,}052\,V_R + 1{,}13{,}58{,}000 = 0$$

Solving this equation, we get, $V_R = 18{,}435$ V

∴ Line voltage at the receiving end = $\sqrt{3} \times 18{,}435 = 31{,}930$ V = **31·93 kV** (*ii*) Line current, $I = \frac{15 \times 10^5}{V_R} = \frac{15 \times 10^5}{18{,}435} = \mathbf{81 \cdot 36\ A}$

(*iii*) Line losses, $= 3 I^2 R = 3 \times (81 \cdot 36)^2 \times 5 \cdot 31 = 1{,}05{,}447\ W = 105 \cdot 447\ kW$

$\therefore$ Transmission efficiency $= \frac{3600}{3600 + 105 \cdot 447} \times 100 = \mathbf{97 \cdot 15\%}$

Example *A short 3-ϕ transmission line with an impedance of $(6 + j\,8)\ \Omega$ per phase has sending and receiving end voltages of 120 kV and 110 kV respectively for some receiving end load at a p.f. of 0·9 lagging. Determine (i) power output and (ii) sending end power factor.*

Solution.

Resistance of each conductor, $R = 6\ \Omega$

Reactance of each conductor, $X_L = 8\ \Omega$

Load power factor, $\cos \phi_R = 0 \cdot 9$ lagging

Receiving end voltage/phase, $V_R = 110 \times 10^3/\sqrt{3} = 63508\ V$

Sending end voltage/phase, $V_S = 120 \times 10^3/\sqrt{3} = 69282\ V$

Let I be the load current. Using approximate expression for V_S, we get,

$$V_S = V_R + I R \cos \phi_R + I X_L \sin \phi_R$$

or $$69282 = 63508 + I \times 6 \times 0 \cdot 9 + I \times 8 \times 0 \cdot 435$$

or $$8 \cdot 88\, I = 5774$$

or $$I = 5774/8 \cdot 88 = 650 \cdot 2\ A$$

(*i*) Power output $= \frac{3\, V_R\, I \cos \phi_R}{1000}\ kW = \frac{3 \times 63508 \times 650 \cdot 2 \times 0 \cdot 9}{1000}$

$= \mathbf{1{,}11{,}490\ kW}$

(*ii*) Sending end p.f., $\cos \phi_S = \frac{V_R \cos \phi_R + I R}{V_S} = \frac{63508 \times 0 \cdot 9 + 650 \cdot 2 \times 6}{69282} = \mathbf{0 \cdot 88\ lag}$

Example *An 11 kV, 3-phase transmission line has a resistance of 1·5 Ω and reactance of 4 Ω per phase. Calculate the percentage regulation and efficiency of the line when a total load of 5000 kVA at 0.8 lagging power factor is supplied at 11 kV at the distant end.*

Solution.

Resistance of each conductor, $R = 1 \cdot 5\ \Omega$

Reactance of each conductor, $X_L = 4\ \Omega$

Receiving end voltage/phase, $V_R = \frac{11 \times 10^3}{\sqrt{3}} = 6351\ V$

Load power factor, $\cos \phi_R = 0 \cdot 8$ lagging

Load current, $I = \frac{\text{Power delivered in kVA} \times 1000}{3 \times V_R}$

$= \frac{5000 \times 1000}{3 \times 6351} = 262 \cdot 43\ A$

Using the approximate expression for V_S (sending end voltage per phase), we get,

$$V_S = V_R + I R \cos \phi_R + I X_L \sin \phi_R$$
$$= 6351 + 262 \cdot 43 \times 1 \cdot 5 \times 0 \cdot 8 + 262 \cdot 43 \times 4 \times 0 \cdot 6 = 7295 \cdot 8\ V$$

% regulation $= \frac{V_S - V_R}{V_R} \times 100 = \frac{7295 \cdot 8 - 6351}{6351} \times 100 = \mathbf{14 \cdot 88\%}$

Line losses $= 3 I^2 R = 3 \times (262 \cdot 43)^2 \times 1 \cdot 5 = 310 \times 10^3\ W = 310\ kW$

$$\text{Output power} = 5000 \times 0{\cdot}8 = 4000 \text{ kW}$$

$$\text{Input power} = \text{Ouput power} + \text{line losses} = 4000 + 310 = 4310 \text{ kW}$$

$$\text{Transmission efficiency} = \frac{\text{Output power}}{\text{Input power}} \times 100 = \frac{4000}{4310} \times 100 = \mathbf{92{\cdot}8\%}$$

Example *A 3-phase, 50 Hz, 16 km long overhead line supplies 1000 kW at 11 kV, 0·8 p.f. lagging. The line resistance is 0·03 Ω per phase per km and line inductance is 0·7 mH per phase per km. Calculate the sending end voltage, voltage regulation and efficiency of transmission.*

Solution.

Resistance of each conductor, $R = 0{\cdot}03 \times 16 = 0{\cdot}48\,\Omega$

Reactance of each conductor, $X_L = 2\pi f L \times 16 = 2\pi \times 50 \times 0{\cdot}7 \times 10^{-3} \times 16 = 3{\cdot}52\,\Omega$

Receiving end voltage/phase, $V_R = \dfrac{11 \times 10^3}{\sqrt{3}} = 6351 \text{ V}$

Load power factor, $\cos\phi_R = 0{\cdot}8$ lagging

$$\text{Line current, } I = \frac{1000 \times 10^3}{3 \times V_R \times \cos\phi} = \frac{1000 \times 10^3}{3 \times 6351 \times 0{\cdot}8} = 65{\cdot}6\text{A}$$

$$\text{Sending end voltage/phase, } V_S = V_R + I R \cos\phi_R + I X_L \sin\phi_R$$
$$= 6351 + 65{\cdot}6 \times 0{\cdot}48 \times 0{\cdot}8 + 65{\cdot}6 \times 3{\cdot}52 \times 0{\cdot}6 = 6515 \text{ V}$$

$$\therefore \quad \text{\%age Voltage regulation} = \frac{V_S - V_R}{V_R} \times 100 = \frac{6515 - 6351}{6351} \times 100 = \mathbf{2{\cdot}58\%}$$

$$\text{Line losses} = 3 I^2 R = 3 \times (65{\cdot}6)^2 \times 0{\cdot}48 = 6{\cdot}2 \times 10^3 \text{ W} = 6{\cdot}2 \text{ kW}$$

$$\text{Input power} = \text{Output power} + \text{Line losses} = 1000 + 6{\cdot}2 = 1006{\cdot}2 \text{ kW}$$

$$\therefore \quad \text{Transmission efficiency} = \frac{\text{Output power}}{\text{Input power}} \times 100 = \frac{1000}{1006{\cdot}2} \times 100 = \mathbf{99{\cdot}38\%}$$

Medium Transmission Lines

In short transmission line calculations, the effects of the line capacitance are neglected because such lines have smaller lengths and transmit power at relatively low voltages (< 20 kV). However, as the length and voltage of the line increase, the capacitance gradually becomes of greater importance. Since medium transmission lines have sufficient length (50-150 km) and usually operate at voltages greater than 20 kV, the effects of capacitance cannot be neglected. Therefore, in order to obtain reasonable accuracy in medium transmission line calculations, the line capacitance must be taken into consideration.

The capacitance is uniformly distributed over the entire length of the line. However, in order to make the calculations simple, the line capacitance is assumed to be lumped or concentrated in the form of capacitors shunted across the line at one or more points. Such a treatment of localising the line capacitance gives reasonably accurate results. The most commonly used methods (known as *localised capacitance methods*) for the solution of medium transmissions lines are :

(*i*) End condenser method (*ii*) Nominal *T* method (*iii*) Nominal π method.

Although the above methods are used for obtaining the performance calculations of medium lines, they can also be used for short lines if their line capacitance is given in a particular problem.

End Condenser Method

In this method, the capacitance of the line is lumped or concentrated at the receiving or load end as shown in Fig. 10.8. This method of localising the line capacitance at the load end overestimates the effects of capacitance. In Fig. 10.8, one phase of the 3-phase transmission line is shown as it is more convenient to work in phase instead of line-to-line values.

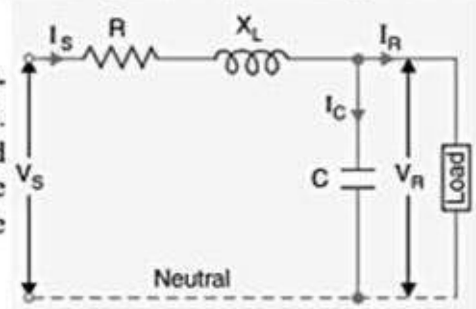


Fig. 10.8

Let I_R = load current per phase

R = resistance per phase

X_L = inductive reactance per phase

C = capacitance per phase

$\cos \phi_R$ = receiving end power factor (*lagging*)

V_S = sending end voltage per phase

The *phasor diagram for the circuit is shown in Fig 10.9. Taking the receiving end voltage $\overrightarrow{V_R}$ as the reference phasor, we have, $\overrightarrow{V_R} = V_R + j\,0$

Load current, $\overrightarrow{I_R} = I_R\,(\cos\phi_R - j\sin\phi_R)$

Capacitive current, $\overrightarrow{I_C} = j\,\overrightarrow{V_R}\,\omega C = j\,2\,\pi f C\,\overrightarrow{V_R}$

The sending end current $\overrightarrow{I_S}$ is the phasor sum of load current $\overrightarrow{I_R}$ and capacitive current $\overrightarrow{I_C}$ *i.e.*,

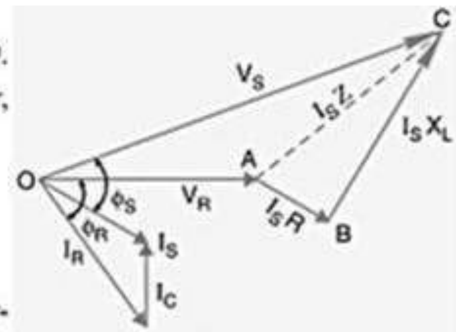


Fig. 10.9

$$
\begin{aligned}
\overrightarrow{I_S} &= \overrightarrow{I_R} + \overrightarrow{I_C}\\
&= I_R\,(\cos\phi_R - j\sin\phi_R) + j\,2\,\pi f C\,V_R\\
&= I_R\cos\phi_R + j\,(-I_R\sin\phi_R + 2\,\pi f C V_R)
\end{aligned}
$$

Voltage drop/phase $= \overrightarrow{I_S}\,\overrightarrow{Z} = \overrightarrow{I_S}\,(R + jX_L)$

Sending end voltage, $\overrightarrow{V_S} = \overrightarrow{V_R} + \overrightarrow{I_S}\,\overrightarrow{Z} = \overrightarrow{V_R} + \overrightarrow{I_S}\,(R + j\,X_L)$

Thus, the magnitude of sending end voltage V_S can be calculated.

$$\%\text{ Voltage regulation} = \frac{V_S - V_R}{V_R} \times 100$$

$$\%\text{ Voltage transmission efficiency} = \frac{\text{Power delivered / phase}}{\text{Power delivered / phase + losses / phase}} \times 100$$

$$= \frac{V_R\,I_R\cos\phi_R}{V_R\,I_R\cos\phi_R + I_S^2\,R} \times 100$$

Limitations. Although end condenser method for the solution of medium lines is simple to work out calculations, yet it has the following drawbacks :

(*i*) There is a considerable error (about 10%) in calculations because the distributed capacitance has been assumed to be lumped or concentrated.

(*ii*) This method overestimates the effects of line capacitance.

Example *A (medium) single phase transmission line 100 km long has the following constants :*

Resistance/km = 0·25 Ω ; *Reactance/km = 0·8 Ω*

Susceptance/km = 14 × 10^{-6} siemen ; *Receiving end line voltage = 66,000 V*

Assuming that the total capacitance of the line is localised at the receiving end alone, determine (i) the sending end current (ii) the sending end voltage (iii) regulation and (iv) supply power factor. The line is delivering 15,000 kW at 0.8 power factor lagging. Draw the phasor diagram to illustrate your calculations.

Solution. Figs. 10.10 (*i*) and (*ii*) show the circuit diagram and phasor diagram of the line respectively.

Total resistance, $R = 0{\cdot}25 \times 100 = 25\,\Omega$

Total reactance, $X_L = 0{\cdot}8 \times 100 = 80\,\Omega$

Total susceptance, $Y = 14 \times 10^{-6} \times 100 = 14 \times 10^{-4}\,S$

Receiving end voltage, $V_R = 66{,}000$ V

$\therefore$ Load current, $I_R = \dfrac{15{,}000 \times 10^3}{66{,}000 \times 0{\cdot}8} = 284$ A

$$\cos\phi_R = 0{\cdot}8\,; \qquad \sin\phi_R = 0{\cdot}6$$

Taking receiving end voltage as the reference phasor [see Fig.10.10 (*ii*)], we have,

$$\overrightarrow{V_R} = V_R + j\,0 = 66{,}000\,V$$

Load current, $\overrightarrow{I_R} = I_R(\cos\phi_R - j\sin\phi_R) = 284\,(0{\cdot}8 - j\,0{\cdot}6) = 227 - j\,170$

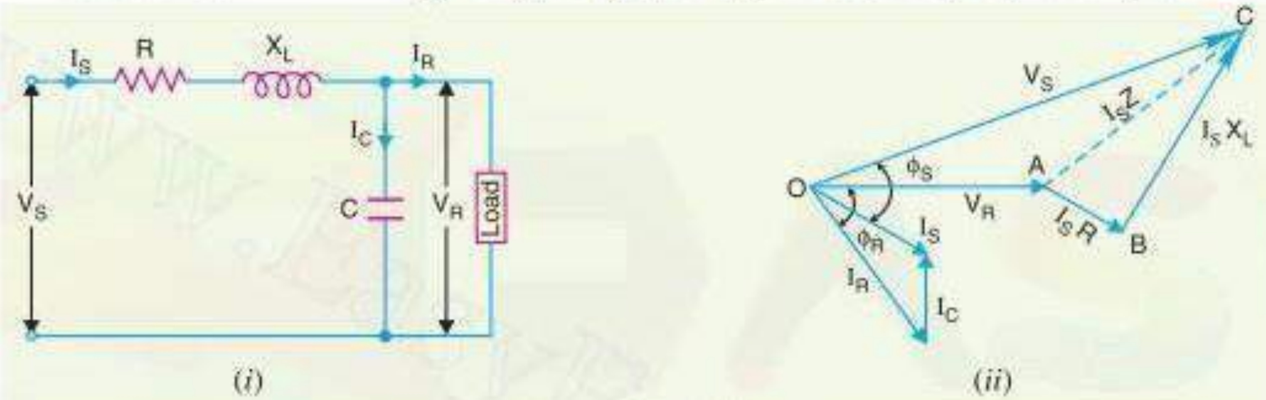


Fig. 10.10

Capacitive current, $\overrightarrow{I_C} = jY \times V_R = j\,14 \times 10^{-4} \times 66000 = j\,92$

(*i*) Sending end current, $\overrightarrow{I_S} = \overrightarrow{I_R} + \overrightarrow{I_C} = (227 - j\,170) + j\,92$

$= 227 - j\,78$ ... (*i*)

Magnitude of $I_S = \sqrt{(227)^2 + (78)^2} =$ **240 A**

(*ii*) Voltage drop $= \overrightarrow{I_S}\,\overrightarrow{Z} = \overrightarrow{I_S}(R + j\,X_L) = (227 - j\,78)\,(25 + j\,80)$

$= 5{,}675 + j\,18{,}160 - j\,1950 + 6240$

$= 11{,}915 + j\,16{,}210$

Sending end voltage, $\overrightarrow{V_S} = \overrightarrow{V_R} + \overrightarrow{I_S}\,\overrightarrow{Z} = 66{,}000 + 11{,}915 + j\,16{,}210$

$= 77{,}915 + j\,16{,}210$ ...(*ii*)

Magnitude of $V_S = \sqrt{(77915)^2 + (16210)^2} =$ **79583V**

(*iii*) % Voltage regulation $= \dfrac{V_S - V_R}{V_R} \times 100 = \dfrac{79{,}583 - 66{,}000}{66{,}000} \times 100 =$ **20·58%**

(*iv*) Referring to exp. (*i*), phase angle between $\overrightarrow{V_R}$ and $\overrightarrow{I_R}$ is :

$$\theta_1 = \tan^{-1} - 78/227 = \tan^{-1}(-0{\cdot}3436) = -18{\cdot}96°$$

Referring to exp. (*ii*), phase angle between $\overrightarrow{V_R}$ and $\overrightarrow{V_S}$ is :

$$\theta_2 = \tan^{-1}\frac{16210}{77915} = \tan^{-1}(0{\cdot}2036) = 11{\cdot}50°$$

$\therefore$ Supply power factor angle, $\phi_S = 18{\cdot}96° + 11{\cdot}50° = 30{\cdot}46°$

$\therefore$ Supply p.f. $= \cos\phi_S = \cos 30{\cdot}46° =$ **0·86 lag**

Nominal T Method

In this method, the whole line capacitance is assumed to be concentrated at the middle point of the line and half the line resistance and reactance are lumped on its either side as shown in Fig. 10.11. Therefore, in this arrangement, full charging current flows over half the line. In Fig. 10.11, one phase of 3-phase transmission line is shown as it is advantageous to work in phase instead of line-to-line values.

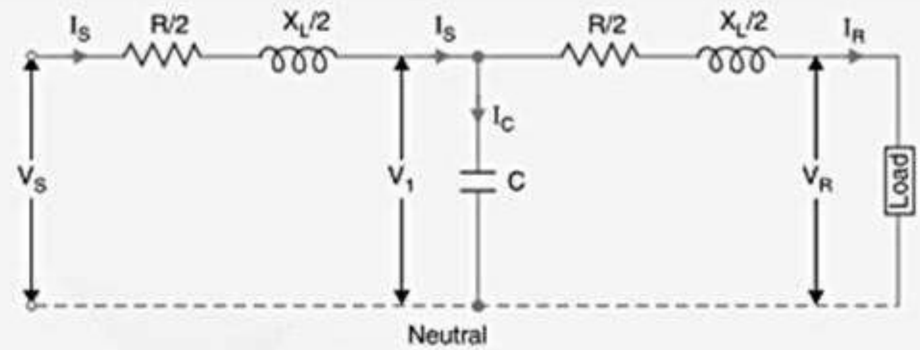


Fig. 10.11

Let
I_R = load current per phase ;
X_L = inductive reactance per phase ;
$\cos \phi_R$ = receiving end power factor (*lagging*) ;
V_1 = voltage across capacitor C
R = resistance per phase
C = capacitance per phase
V_S = sending end voltage/phase

The *phasor diagram for the circuit is shown in Fig. 10.12. Taking the receiving end voltage $\overrightarrow{V_R}$ as the reference phasor, we have,

Receiving end voltage, $\overrightarrow{V_R} = V_R + j\,0$

Load current, $\overrightarrow{I_R} = I_R(\cos \phi_R - j \sin \phi_R)$

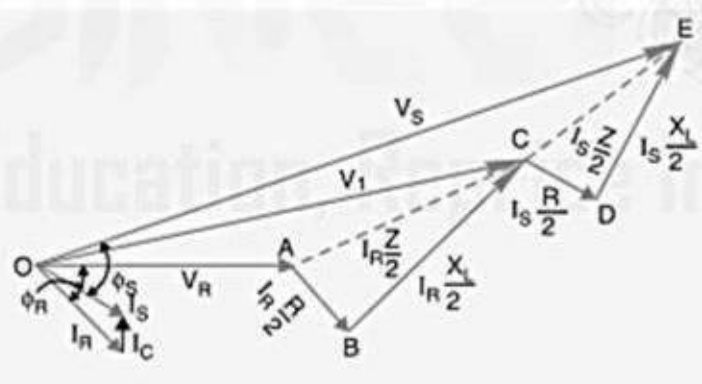


Fig. 10.12

Voltage across C, $\quad \vec{V_1} = \vec{V_R} + \vec{I_R}\,\vec{Z}/2$

$$= V_R + I_R(\cos\phi_R - j\sin\phi_R)\left(\frac{R}{2} + j\frac{X_L}{2}\right)$$

Capacitive current, $\quad \vec{I_C} = j\,\omega\,C\,\vec{V_1} = j\,2\pi f C\,\vec{V_1}$

Sending end current, $\quad \vec{I_S} = \vec{I_R} + \vec{I_C}$

Sending end voltage, $\quad \vec{V_S} = \vec{V_1} + \vec{I_S}\,\frac{\vec{Z}}{2} = \vec{V_1} + \vec{I_S}\left(\frac{R}{2} + j\frac{X_L}{2}\right)$

Example *A 3-phase, 50-Hz overhead transmission line 100 km long has the following constants :*

Resistance/km/phase $= 0.1\ \Omega$

Inductive reactance/km/phase $= 0{\cdot}2\ \Omega$

Capacitive susceptance/km/phase $= 0{\cdot}04 \times 10^{-4}$ *siemen*

Determine (i) the sending end current (ii) sending end voltage (iii) sending end power factor and (iv) transmission efficiency when supplying a balanced load of 10,000 kW at 66 kV, p.f. 0·8 lagging. Use nominal T method.

Solution. Figs. 10.13 (*i*) and 10.13 (*ii*) show the circuit diagram and phasor diagram of the line respectively.

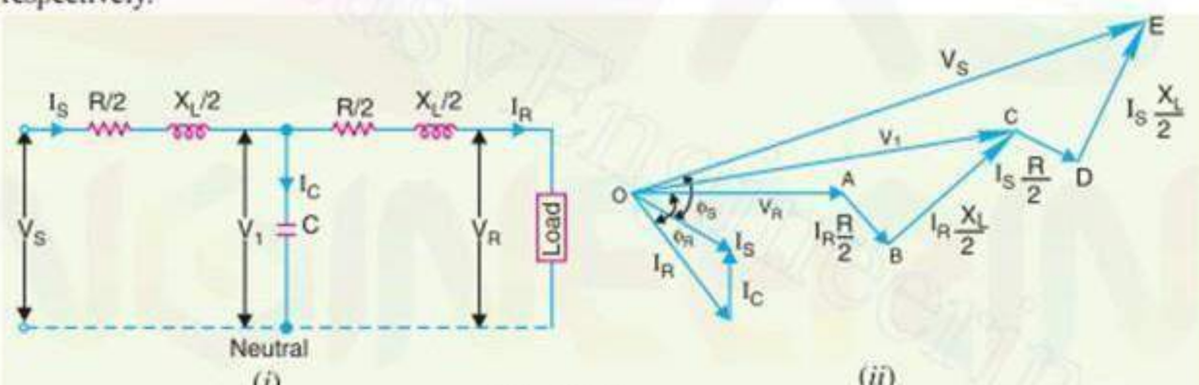


Fig. 10.13

Total resistance/phase, $\quad R = 0{\cdot}1 \times 100 = 10\ \Omega$

Total reactance/phase. $\quad X_L = 0{\cdot}2 \times 100 = 20\ \Omega$

Capacitive susceptance, $\quad Y = 0{\cdot}04 \times 10^{-4} \times 100 = 4 \times 10^{-4}$ S

Receiving end voltage/phase, $\quad V_R = 66{,}000/\sqrt{3} = 38105\ V$

Load current, $\quad I_R = \dfrac{10{,}000 \times 10^3}{\sqrt{3} \times 66 \times 10^3 \times 0{\cdot}8} = 109$ A

$$\cos\phi_R = 0{\cdot}8\ ;\ \ \sin\phi_R = 0{\cdot}6$$

Impedance per phase, $\quad \vec{Z} = R + j\,X_L = 10 + j\,20$

(*i*) Taking receiving end voltage as the reference phasor [see Fig. 10.13 (*ii*)], we have,

Receiving end voltage, $\quad \vec{V_R} = V_R + j\,0 = 38{,}105\ V$

Load current, $\quad \vec{I_R} = I_R(\cos\phi_R - j\sin\phi_R) = 109\,(0{\cdot}8 - j\,0{\cdot}6) = 87{\cdot}2 - j\,65{\cdot}4$

Voltage across C, $\quad \vec{V_1} = \vec{V_R} + \vec{I_R}\,\vec{Z}/2 = 38,105 + (87{\cdot}2 - j\,65{\cdot}4)\,(5 + j\,10)$

$$= 38{,}105 + 436 + j\,872 - j\,327 + 654 = 39{,}195 + j\,545$$

Charging current, $\overrightarrow{I_C} = jY\overrightarrow{V_1} = j\,4\times10^{-4}(39{,}195 + j\,545) = -0\cdot218 + j\,15\cdot6$

Sending end current, $\overrightarrow{I_S} = \overrightarrow{I_R} + \overrightarrow{I_C} = (87\cdot2 - j\,65\cdot4) + (-0\cdot218 + j\,15\cdot6)$

$= 87\cdot0 - j\,49\cdot8 = 100\angle -29°47'$ A

$\therefore$ Sending end current $=$ **100 A**

(*ii*) Sending end voltage, $\overrightarrow{V_S} = \overrightarrow{V_1} + \overrightarrow{I_S}\,\overrightarrow{Z}/2 = (39{,}195 + j\,545) + (87\cdot0 - j\,49\cdot8)\,(5 + j\,10)$

$= 39{,}195 + j\,545 + 434\cdot9 + j\,870 - j\,249 + 498$

$= 40128 + j\,1170 = 40145\angle 1°40'$ V

$\therefore$ Line value of sending end voltage

$= 40145\times\sqrt{3} = 69\,533$ V $=$ **69·533 kV**

(*iii*) Referring to phasor diagram in Fig. 10.14,

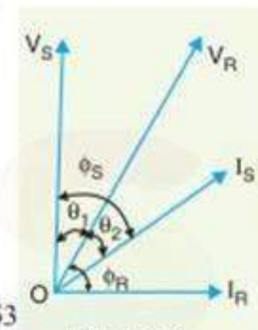


Fig. 10.14

θ_1 = angle between $\overrightarrow{V_R}$ and $\overrightarrow{V_S} = 1°40'$

θ_2 = angle between $\overrightarrow{V_R}$ and $\overrightarrow{I_S} = 29°\,47'$

$\therefore$ ϕ_S = angle between $\overrightarrow{V_S}$ and $\overrightarrow{I_S}$

$= \theta_1 + \theta_2 = 1°40' + 29°47' = 31°27'$

$\therefore$ Sending end power factor, $\cos\phi_S = \cos 31°27' =$ **0·853 lag**

(*iv*) Sending end power $= 3\,V_S I_S\cos\phi_S = 3\times40{,}145\times100\times0\cdot853$

$= 10273105$ W $= 10273\cdot105$ kW

Power delivered $= 10{,}000$ kW

$\therefore$ Transmission efficiency $= \dfrac{10{,}000}{10273\cdot105}\times100 =$ **97·34%**

Example *A 3-phase, 50 Hz transmission line 100 km long delivers 20 MW at 0·9 p.f. lagging and at 110 kV. The resistance and reactance of the line per phase per km are 0·2 Ω and 0·4 Ω respectively, while capacitance admittance is 2·5 × 10^{-6} siemen/km/phase. Calculate :* (*i*) *the current and voltage at the sending end* (*ii*) *efficiency of transmission. Use nominal T method.*

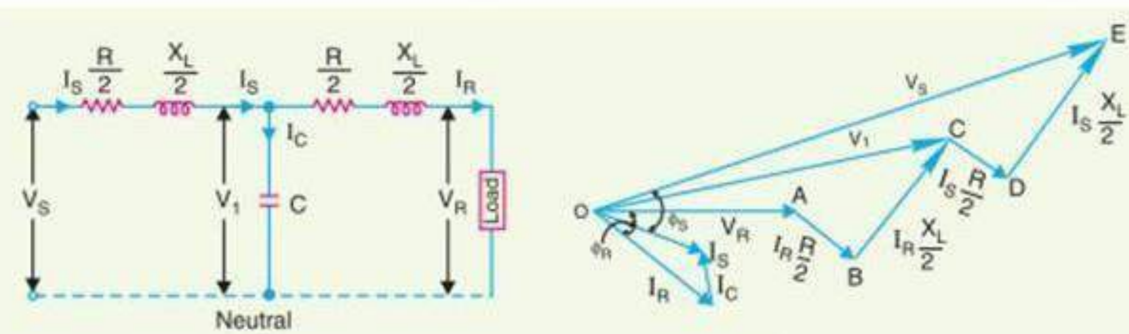
Solution. Figs. 10.15 (*i*) and 10.15 (*ii*) show the circuit diagram and phasor diagram respectively.

Total resistance/phase, $R = 0\cdot2\times100 = 20\ \Omega$

Total reactance/phase, $X_L = 0\cdot4\times100 = 40\ \Omega$

Total capacitance admittance/phase, $Y = 2\cdot5\times10^{-6}\times100 = 2\cdot5\times10^{-4}$ S

Phase impedance, $\overrightarrow{Z} = 20 + j40$



(*i*) (*ii*)

Fig. 10.15

Receiving end voltage/phase, $V_R = 110 \times 10^3/\sqrt{3} = 63508$ V

$$\text{Load current, } I_R = \frac{20 \times 10^6}{\sqrt{3} \times 110 \times 10^3 \times 0{\cdot}9} = 116.6 \text{ A}$$

$$\cos \phi_R = 0{\cdot}9 \; ; \sin \phi_R = 0{\cdot}435$$

(*i*) Taking receiving end voltage as the reference phasor [see phasor diagram 10.15 (*ii*)], we have,

$$\vec{V_R} = V_R + j0 = 63508 \text{ V}$$

Load current, $\vec{I_R} = I_R(\cos \phi_R - j \sin \phi_R) = 116{\cdot}6\,(0{\cdot}9 - j\,0{\cdot}435) = 105 - j50{\cdot}7$

Voltage across C, $\vec{V_1} = \vec{V_R} + \vec{I_R}\,\vec{Z}/2 = 63508 + (105 - j\,50{\cdot}7)(10 + j\,20)$

$$= 63508 + (2064 + j1593) = 65572 + j1593$$

Charging current, $\vec{I_C} = j\,Y\vec{V_1} = j\,2{\cdot}5 \times 10^{-4}\,(65572 + j1593) = -0{\cdot}4 + j\,16{\cdot}4$

Sending end current, $\vec{I_S} = \vec{I_R} + \vec{I_C} = (105 - j\,50{\cdot}7) + (-0{\cdot}4 + j16{\cdot}4)$

$$= (104{\cdot}6 - j\,34{\cdot}3) = 110 \angle -18°9' \text{ A}$$

$\therefore$ Sending end current $= 110$ A

Sending end voltage, $\vec{V_S} = \vec{V_1} + \vec{I_S}\,\vec{Z}/2$

$$= (65572 + j1593) + (104{\cdot}6 - j34{\cdot}3)(10 + j20)$$

$$= 67304 + j\,3342$$

$\therefore$ Magnitude of $V_S = \sqrt{(67304)^2 + (3342)^2} = 67387\ V$

$\therefore$ Line value of sending end voltage

$$= 67387 \times \sqrt{3} = 116717 \text{ V} = 116{\cdot}717 \text{ kV}$$

(*ii*) Total line losses for the three phases

$$= 3\,I_S^2\,R/2 + 3I_R^2\,R/2$$

$$= 3 \times (110)^2 \times 10 + 3 \times (116{\cdot}6)^2 \times 10$$

$$= 0{\cdot}770 \times 10^6 \text{ W} = 0{\cdot}770 \text{ MW}$$

$\therefore$ Transmission efficiency $= \dfrac{20}{20 + 0{\cdot}770} \times 100 = 96{\cdot}29\%$

Nominal π Method

In this method, capacitance of each conductor (*i.e.*, line to neutral) is divided into two halves; one half being lumped at the sending end and the other half at the receiving end as shown in Fig. 10.16. It is obvious that capacitance at the sending end has no effect on the line drop. However, its charging current must be added to line current in order to obtain the total sending end current.

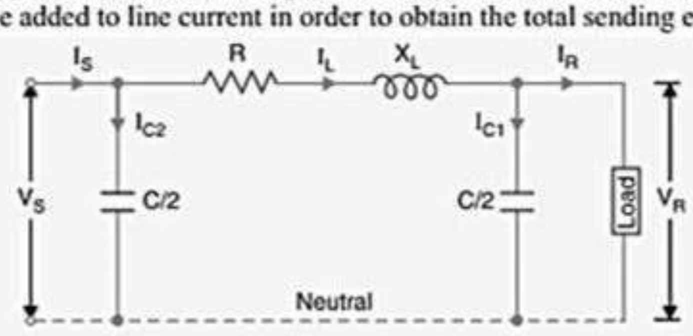


Fig. 10.16

Let $\quad I_R$ = load current per phase

R = resistance per phase

X_L = inductive reactance per phase

C = capacitance per phase

$\cos \phi_R$ = receiving end power factor (*lagging*)

V_S = sending end voltage per phase

The *phasor diagram for the circuit is shown in Fig. 10.17. Taking the receiving end voltage as the reference phasor, we have,

$$\overrightarrow{V_R} = V_R + j0$$

Load current, $\quad \overrightarrow{I_R} = I_R(\cos \phi_R - j \sin \phi_R)$

Charging current at load end is

$$\overrightarrow{I_{C1}} = j\,\omega\,(C/2)\,\overrightarrow{V_R} = j\,\pi\,f\,C\,\overrightarrow{V_R}$$

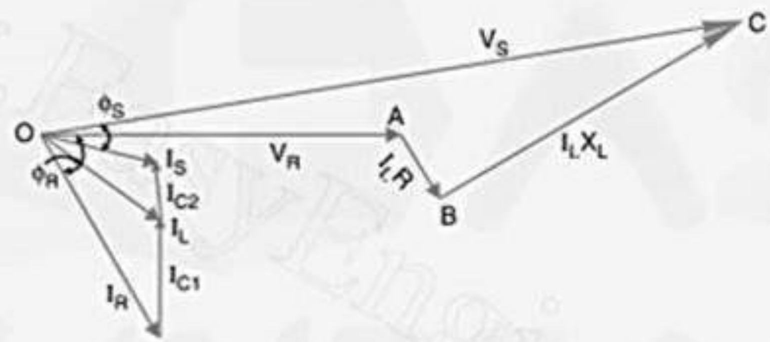


Fig. 10.17

Line current, $\quad \overrightarrow{I_L} = \overrightarrow{I_R} + \overrightarrow{I_{C1}}$

Sending end voltage, $\quad \overrightarrow{V_S} = \overrightarrow{V_R} + \overrightarrow{I_L}\,\overrightarrow{Z} = \overrightarrow{V_R} + \overrightarrow{I_L}\,(R + jX_L)$

Charging current at the sending end is

$$\overrightarrow{I_{C2}} = j\,\omega\,(C/2)\,\overrightarrow{V_S} = j\,\pi\,f\,C\,\overrightarrow{V_S}$$

$\therefore$ Sending end current, $\quad \overrightarrow{I_S} = \overrightarrow{I_L} + \overrightarrow{I_{C2}}$

Example *A 3-phase, 50Hz, 150 km line has a resistance, inductive reactance and capacitive shunt admittance of 0·1 Ω, 0·5 Ω and 3×10^{-6} S per km per phase. If the line delivers 50 MW at 110 kV and 0·8 p.f. lagging, determine the sending end voltage and current. Assume a nominal π circuit for the line.*

Solution. Fig. 10.18 shows the circuit diagram for the line.

Total resistance/phase, $\quad R = 0{\cdot}1 \times 150 = 15\ \Omega$

Total reactance/phase, $\quad X_L = 0{\cdot}5 \times 150 = 75\ \Omega$

Capacitive admittance/phase, $\quad Y = 3 \times 10^{-6} \times 150 = 45 \times 10^{-5}\ \text{S}$

Receiving end voltage/phase, $\quad V_R = 110 \times 10^3/\sqrt{3} = 63{,}508\ \text{V}$

Load current, $\quad I_R = \dfrac{50 \times 10^6}{\sqrt{3} \times 110 \times 10^3 \times 0{\cdot}8} = 328\ \text{A}$

$$\cos\phi_R = 0{\cdot}8\ ;\quad \sin\phi_R = 0{\cdot}6$$

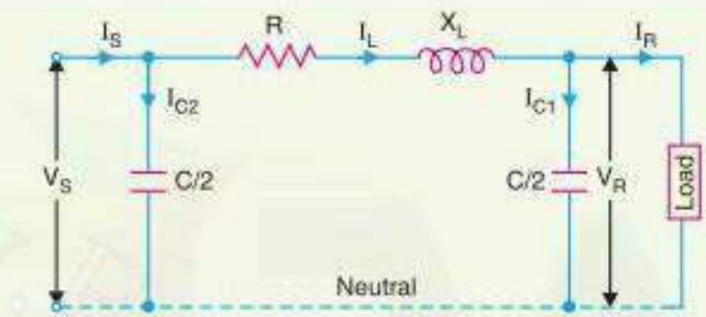


Fig. 10.18

Taking receiving end voltage as the reference phasor, we have,

$$\overrightarrow{V_R} = V_R + j\,0 = 63{,}508\ \text{V}$$

Load current, $\quad \overrightarrow{I_R} = I_R(\cos\phi_R - j\sin\phi_R) = 328\,(0{\cdot}8 - j0{\cdot}6) = 262{\cdot}4 - j196{\cdot}8$

Charging current at the load end is

$$\overrightarrow{I_{C1}} = \overrightarrow{V_R}\, j\,\frac{Y}{2} = 63{,}508 \times j\,\frac{45 \times 10^{-5}}{2} = j\,14{\cdot}3$$

Line current, $\quad \overrightarrow{I_L} = \overrightarrow{I_R} + \overrightarrow{I_{C1}} = (262{\cdot}4 - j\,196{\cdot}8) + j\,14{\cdot}3 = 262{\cdot}4 - j\,182{\cdot}5$

Sending end voltage, $\quad \overrightarrow{V_S} = \overrightarrow{V_R} + \overrightarrow{I_L}\,\vec{Z} = \overrightarrow{V_R} + \overrightarrow{I_L}\,(R + j\,X_L)$

$$= 63{,}508 + (262{\cdot}4 - j\,182{\cdot}5)(15 + j\,75)$$
$$= 63{,}508 + 3936 + j\,19{,}680 - j\,2737{\cdot}5 + 13{,}687$$
$$= 81{,}131 + j\,16{,}942{\cdot}5 = 82{,}881\ \angle\, 11°47'\ \text{V}$$

$\therefore$ Line to line sending end voltage $= 82{,}881 \times \sqrt{3} = 1{,}43{,}550\ \text{V} =$ **143·55 kV**

Charging current at the sending end is

$$I_{C2} = j\overrightarrow{V_S}\,Y/2 = (81{,}131 + j\,16{,}942{\cdot}5)\,j\,\frac{45 \times 10^{-5}}{2}$$
$$= -3{\cdot}81 + j\,18{\cdot}25$$

Sending end current, $\quad \overrightarrow{I_S} = \overrightarrow{I_L} + \overrightarrow{I_{C_2}} = (262{\cdot}4 - j\,182{\cdot}5) + (-3{\cdot}81 + j\,18{\cdot}25)$

$$= 258{\cdot}6 - j\,164{\cdot}25 = 306{\cdot}4\angle -32{\cdot}4°\ \text{A}$$

$\therefore$ Sending end current $=$ **306·4 A**

Chapter5 EHV transmission

EHV AC Transmission:

NECESSITY OF EHVAC TRANSMISSION:

1. With the increase in transmission voltage, for same amount of power to be transmitted current in the line decreases which reduces I2R losses. This will lead to increase in transmission efficiency.

2. With decrease in transmission current, size of conductor required reduces which decreases the volume of conductor.

3. The transmission capacity is proportional to square of operating voltages. Thus the transmission capacity of line increases with increase in voltage.

4. With increase in level of transmission voltage, the installation cost of the transmission line per km decreases.

5. It is economical with EHV transmission to interconnect the power systems on a large scale.

6. The no. of circuits and the land requirement for transmission decreases with the use of higher transmission voltages.

ADVANTAGES :

- Reduction in the current.
- Reduction in the losses.
- Reduction in volume of conductor material required.
- Decrease in voltage drop & improvement of voltage regulation.
- Increase in Transmission Efficiency.
- Increased power handling capacity.
- The no. of circuits & the land requirement reduces as transmission voltage increases.
- The total line cost per MW per km decreases considerably with the increase in line voltage.

PROBLEMS INVOLVED IN EHV TRANSMISSION:

1. Corona loss and radio interference
2. Heavy supporting structure and erection difficulties
3. Insulation requirement
4. Suitability considerations
5. Current carrying capacity
6. Ferranti effect
7. Environmental and biological aspects
8. Equipment cost

HVDC Transmission System

We know that AC power is generated in the generating station. This should first be converted into DC. The conversion is done with the help of rectifier. The DC power will flow through the overhead lines. At the user end, this DC has to be converted into AC. For that purpose, an inverter is placed at the receiving end.

Thus, there will be a rectifier terminal in one end of HVDC substation and an inverter terminal in the other end. The power of the sending end and user end will be always equal (Input Power = Output Power).

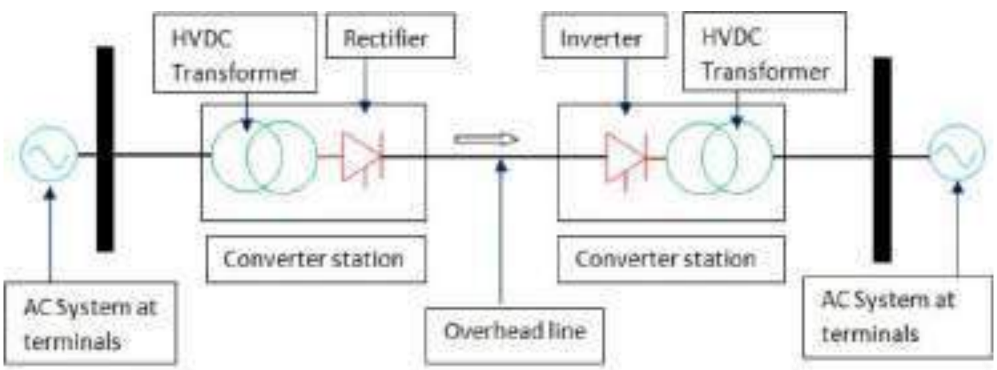


Figure 1: HVDC Substation Layout

When there are two converter stations at both ends and a single transmission line is termed as two terminal DC systems. When there are two or more converter stations and DC transmission lines is termed as multi-terminal DC substation.

Comparison of both HVAC and HVDC Transmission System

HVDC Transmission System	HVAC Transmission System
Low losses.	Losses are high due to the skin effect and corona discharge

Better Voltage regulation and Control ability.	Voltage regulation and Control ability is low.
Transmit more power over a longer distance.	Transmit less power compared to a HVDC system.
Less insulation is needed.	More insulation is required.
Reliability is high.	Low Reliability.
Asynchronous interconnection is possible.	Asynchronous interconnection is not possible.
Reduced line cost due to fewer conductors.	Line cost is high.
Towers are cheaper, simple and narrow.	Towers are bigger compared to HVDC.

Disadvantages of HVDC Transmission

- Converters with small overload capacity are used.
- Circuit Breakers, Converters and AC filters are expensive especially for small distance transmission.
- No transformers for altering the voltage level.
- HVDC link is extremely complicated.
- Uncontrollable power flow.

Application of HVDC Transmission

- Undersea and underground cables
- AC network interconnections
- Interconnecting Asynchronous system

Chapter 6 Distribution systems

Types of D.C. Distributors

The most general method of classifying d.c. distributors is the way they are fed by the feeders. On this basis, d.c. distributors are classified as:

(*i*) Distributor fed at one end
(*ii*) Distributor fed at both ends
(*iii*) Distributor fed at the centre
(*iv*) Ring distributor.

(*i*) **Distributor fed at one end.** In this type of feeding, the distributor is connected to the supply at one end and loads are taken at different points along the length of the distributor. Fig. 13.1 shows the single line diagram of a d.c. distributor *A B* fed at the end *A* (also known as *singly fed distributor*) and loads I_1, I_2 and I_3 tapped off at points *C*, *D* and *E* respectively.

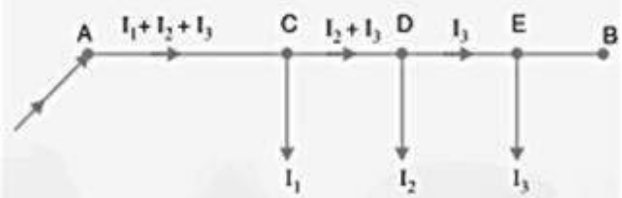


Fig. 13.1

The following points are worth noting in a singly fed distributor :

(*a*) The current in the various sections of the distributor away from feeding point goes on decreasing. Thus current in section *AC* is more than the current in section *CD* and current in section *CD* is more than the current in section *DE*.

(*b*) The voltage across the loads away from the feeding point goes on decreasing. Thus in Fig. 13.1, the minimum voltage occurs at the load point *E*.

(*c*) In case a fault occurs on any section of the distributor, the whole distributor will have to be disconnected from the supply mains. Therefore, continuity of supply is interrupted.

(*ii*) **Distributor fed at both ends.** In this type of feeding, the distributor is connected to the supply mains at both ends and loads are tapped off at different points along the length of the distributor. The voltage at the feeding points may or may not be equal. Fig. 13.2 shows a distributor *A B* fed at the ends *A* and *B* and loads of I_1, I_2 and I_3 tapped off at points *C*, *D* and *E* respectively. Here, the load voltage goes on decreasing as we move away from one feeding point *say A*, reaches minimum value and then again starts rising and reaches maximum value when we reach the other feeding point *B*. The minimum voltage occurs at some load point and is never fixed. It is shifted with the variation of load on different sections of the distributor.

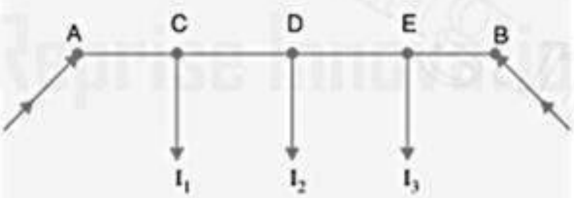


Fig. 13.2

Advantages

(*a*) If a fault occurs on any feeding point of the distributor, the continuity of supply is maintained from the other feeding point.

(*b*) In case of fault on any section of the distributor, the continuity of supply is maintained from the other feeding point.

(c) The area of X-section required for a doubly fed distributor is much less than that of a singly fed distributor.

(iii) **Distributor fed at the centre.** In this type of feeding, the centre of the distributor is connected to the supply mains as shown in Fig. 13.3. It is equivalent to two singly fed distributors, each distributor having a common feeding point and length equal to half of the total length.

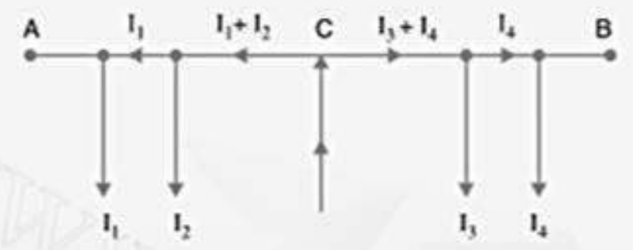


Fig. 13.3

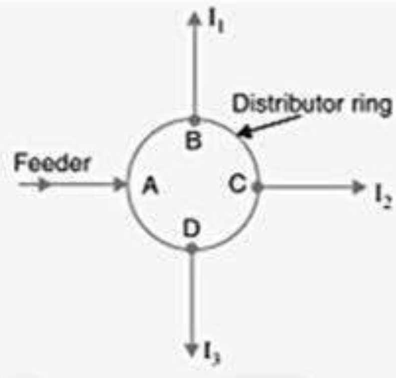
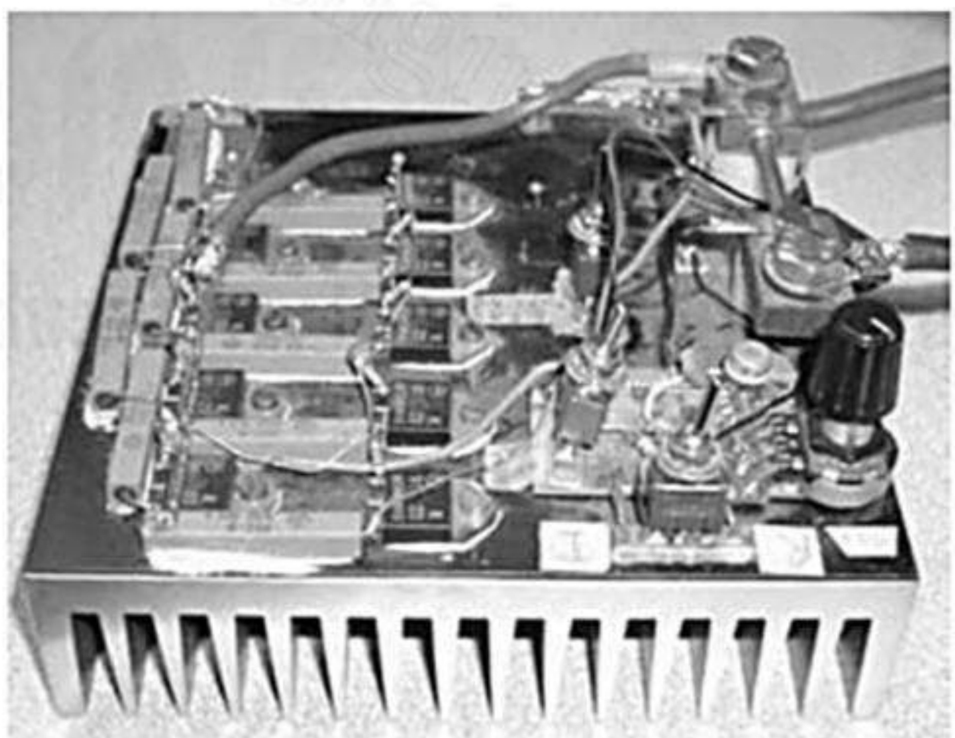


Fig. 13.4

(iv) **Ring mains.** In this type, the distributor is in the form of a closed ring as shown in Fig.13.4. It is equivalent to a straight distributor fed at both ends with equal voltages, the two ends being brought together to form a closed ring. The distributor ring may be fed at one or more than one point.

D.C. Distribution Calculations

In addition to the methods of feeding discussed above, a distributor may have (i) concentrated loading (ii) uniform loading (iii) both concentrated and uniform loading. The concentrated loads are those which act on particular points of the distributor. A common example of such loads is that tapped off for domestic use. On the other hand, distributed loads are those which act uniformly on all points of the distributor.



D.C. Load

Ideally, there are no distributed loads. However, a nearest example of distributed load is a large number of loads of same wattage connected to the distributor at equal distances.

In d.c. distribution calculations, one important point of interest is the determination of point of minimum potential on the distributor. The point where it occurs depends upon the loading conditions and the method of feeding the distributor. The distributor is so designed that the minimum potential on it is not less than 6% of rated voltage at the consumer's terminals. In the next sections, we shall discuss some important cases of d.c. distributors separately.

D.C. Distributor Fed at one End—Concentrated Loading

Fig. 13.5 shows the single line diagram of a 2-wire d.c. distributor *A B* fed at one end *A* and having concentrated loads I_1, I_2, I_3 and I_4 tapped off at points *C, D, E* and *F* respectively.

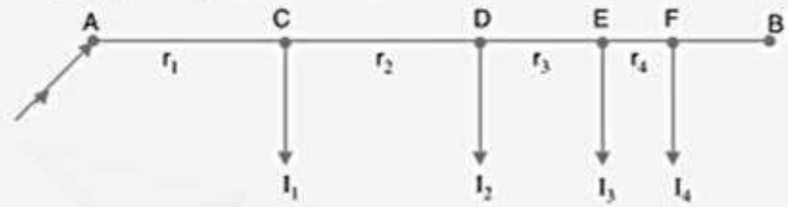


Fig. 13.5

Let r_1, r_2, r_3 and r_4 be the resistances of both wires (go and return) of the sections *A C, CD, DE* and *EF* of the distributor respectively.

Current fed from point *A* $= I_1 + I_2 + I_3 + I_4$

Current in section *A C* $= I_1 + I_2 + I_3 + I_4$

Current in section *CD* $= I_2 + I_3 + I_4$

Current in section *DE* $= I_3 + I_4$

Current in section *EF* $= I_4$

Voltage drop in section *A C* $= r_1 (I_1 + I_2 + I_3 + I_4)$

Voltage drop in section *CD* $= r_2 (I_2 + I_3 + I_4)$

Voltage drop in section *DE* $= r_3 (I_3 + I_4)$

Voltage drop in section *EF* $= r_4 I_4$

∴ Total voltage drop in the distributor

$$= r_1 (I_1 + I_2 + I_3 + I_4) + r_2 (I_2 + I_3 + I_4) + r_3 (I_3 + I_4) + r_4 I_4$$

It is easy to see that the minimum potential will occur at point *F* which is farthest from the feeding point *A*.

Example . *A 2-wire d.c. distributor cable AB is 2 km long and supplies loads of 100A, 150A, 200A and 50A situated 500 m, 1000 m, 1600 m and 2000 m from the feeding point A. Each conductor has a resistance of 0·01 Ω per 1000 m. Calculate the p.d. at each load point if a p.d. of 300 V is maintained at point A.*

Solution. Fig. 13.6 shows the single line diagram of the distributor with its tapped currents.

Resistance per 1000 m of distributor $= 2 \times 0{\cdot}01 = 0{\cdot}02\,\Omega$

Resistance of section $AC, R_{AC} = 0{\cdot}02 \times 500/1000 = 0{\cdot}01\ \Omega$

Resistance of section $CD, R_{CD} = 0{\cdot}02 \times 500/1000 = 0{\cdot}01\ \Omega$

Resistance of section $DE, R_{DE} = 0{\cdot}02 \times 600/1000 = 0{\cdot}012\ \Omega$

Resistance of section $EB, R_{EB} = 0{\cdot}02 \times 400/1000 = 0{\cdot}008\ \Omega$

Referring to Fig. 13.6, the currents in the various sections of the distributor are :

$I_{EB} = 50$ A ; $I_{DE} = 50 + 200 = 250$ A

$I_{CD} = 250 + 150 = 400$ A ; $I_{AC} = 400 + 100 = 500$ A

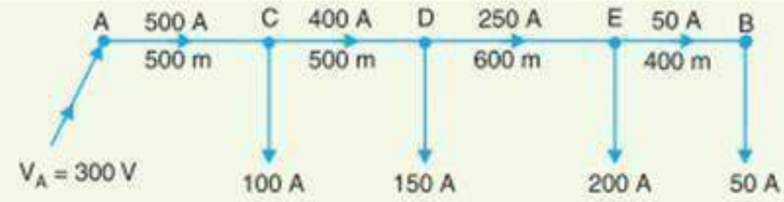


Fig. 13.6

P.D. at load point C, $\quad V_C$ = Voltage at A − Voltage drop in AC

$$= V_A - I_{AC} R_{AC}$$

$$= 300 - 500 \times 0{\cdot}01 = \mathbf{295\ V}$$

P.D. at load point D, $\quad V_D = V_C - I_{CD} R_{CD}$

$$= 295 - 400 \times 0{\cdot}01 = \mathbf{291\ V}$$

P.D. at load point E, $\quad V_E = V_D - I_{DE} R_{DE}$

$$= 291 - 250 \times 0{\cdot}012 = \mathbf{288\ V}$$

P.D. at load point B, $\quad V_B = V_E - I_{EB} R_{EB}$

$$= 288 - 50 \times 0{\cdot}008 = \mathbf{287{\cdot}6\ V}$$

Example *A 2-wire d.c. distributor AB is 300 metres long. It is fed at point A. The various loads and their positions are given below :*

At point	*distance from A in metres*	*concentrated load in amperes*
C	*40*	*30*
D	*100*	*40*
E	*150*	*100*
F	*250*	*50*

If the maximum permissible voltage drop is not to exceed 10 V, find the cross-sectional area of the distributor. Take $\rho = 1{\cdot}78 \times 10^{-8}\ \Omega m$.

Solution. The single line diagram of the distributor along with its tapped currents is shown in Fig. 13.7. Suppose that resistance of 100 m length of the distributor is r ohms. Then resistance of various sections of the distributor is :

$$R_{AC} = 0{\cdot}4\,r\,\Omega \;;\; R_{CD} = 0{\cdot}6\,r\,\Omega \;;\; R_{DE} = 0{\cdot}5\,r\,\Omega \;;\; R_{EF} = r\,\Omega$$

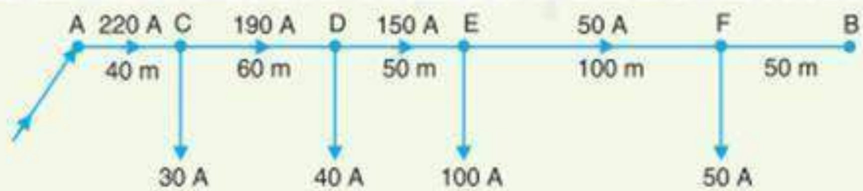


Fig. 13.7

Referring to Fig. 13.7, the currents in the various sections of the distributor are :

$$I_{AC} = 220\text{ A} \;;\; I_{CD} = 190\text{ A} \;;\; I_{DE} = 150\text{ A} \;;\; I_{EF} = 50\text{ A}$$

Total voltage drop over the distributor

$$= I_{AC} R_{AC} + I_{CD} R_{CD} + I_{DE} R_{DE} + I_{EF} R_{EF}$$

$$= 220 \times 0{\cdot}4r + 190 \times 0{\cdot}6r + 150 \times 0{\cdot}5r + 50 \times r$$

$$= 327\ r$$

As the maximum permissible drop in the distributor is 10 V,

$$\therefore \qquad 10 = 327\ r$$

or $\qquad r = 10/327 = 0{\cdot}03058\ \Omega$

$$\text{X-sectional area of conductor} = \frac{*\rho l}{r/2} = \frac{1{\cdot}78\times10^{-8}\times100}{\frac{0{\cdot}03058}{2}} = 116{\cdot}4\times10^{-6}\ \text{m}^2 = \mathbf{1{\cdot}164\ cm^2}$$

Example *Two tram cars (A & B) 2 km and 6 km away from a sub-station return 40 A and 20 A respectively to the rails. The sub-station voltage is 600 V d.c. The resistance of trolley wire is 0·25 Ω/km and that of track is 0·03 Ω/km. Calculate the voltage across each tram car.*

Solution. The tram car operates on d.c. supply. The positive wire is placed overhead while the rail track acts as the negative wire. Fig. 13.8 shows the single line diagram of the arrangement.

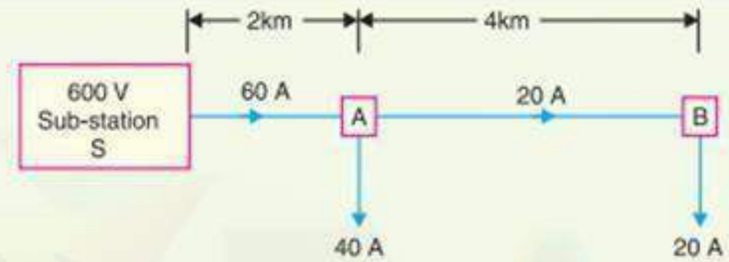


Fig. 13.8

Resistance of trolley wire and track/km

$$= 0{\cdot}25 + 0{\cdot}03 = 0{\cdot}28\Omega$$

$$\text{Current in section } SA = 40 + 20 = 60 \text{ A}$$

$$\text{Current in section } AB = 20 \text{ A}$$

$$\text{Voltage drop in section } SA = 60\times0{\cdot}28\times2 = 33{\cdot}6 \text{ V}$$

$$\text{Voltage drop in section } AB = 20\times0{\cdot}28\times4 = 22{\cdot}4 \text{ V}$$

$$\therefore \quad \text{Voltage across tram } A = 600 - 33{\cdot}6 = \mathbf{566{\cdot}4\ V}$$

$$\text{Voltage across tram } B = 566{\cdot}4 - 22{\cdot}4 = \mathbf{544\ V}$$

Uniformly Loaded Distributor Fed at One End

Fig 13.11 shows the single line diagram of a 2-wire d.c. distributor AB fed at one end A and loaded uniformly with i amperes per metre length. It means that at every 1 m length of the distributor, the load tapped is i amperes. Let l metres be the length of the distributor and r ohm be the resistance per metre run.

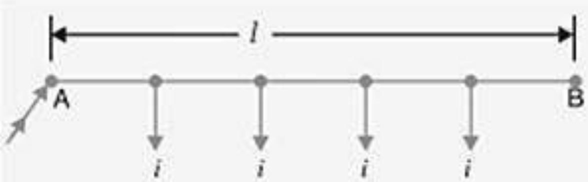


Fig. 13.11

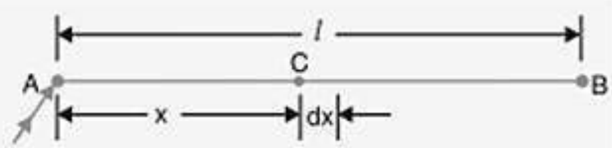


Fig. 13.12

Consider a point C on the distributor at a distance x metres from the feeding point A as shown in Fig. 13.12. Then current at point C is

$$= i\,l - i\,x \text{ amperes} = i\,(l - x) \text{ amperes}$$

Now, consider a small length dx near point C. Its resistance is $r\ dx$ and the voltage drop over length dx is

$$dv = i(l-x)\,r\,dx = i\,r\,(l-x)\,dx$$

Total voltage drop in the distributor upto point C is

$$v = \int_0^x i\,r\,(l-x)\,dx = i\,r\left(l\,x - \frac{x^2}{2}\right)$$

The voltage drop upto point B (*i.e.* over the whole distributor) can be obtained by putting $x = l$ in the above expression.

∴ Voltage drop over the distributor AB

$$= i\,r\left(l \times l - \frac{l^2}{2}\right)$$

$$= \tfrac{1}{2}\,i\,r\,l^2 = \tfrac{1}{2}\,(i\,l)\,(r\,l)$$

$$= \tfrac{1}{2}\,I\,R$$

where $i\,l = I$, the total current entering at point A

$r\,l = R$, the total resistance of the distributor

Thus, in a uniformly loaded distributor fed at one end, the total voltage drop is equal to that produced by the whole of the load assumed to be concentrated at the middle point.

Example *A 2-wire d.c. distributor 200 metres long is uniformly loaded with 2A/metre. Resistance of single wire is 0·3 Ω/km. If the distributor is fed at one end, calculate :*

(*i*) *the voltage drop upto a distance of 150 m from the feeding point*

(*ii*) *the maximum voltage drop*

Solution.

Current loading, $i = 2$ A/m

Resistance of distributor per metre run,

$$r = 2 \times 0{\cdot}3/1000 = 0{\cdot}0006\,\Omega$$

Length of distributor, $l = 200$ m

(*i*) Voltage drop upto a distance x metres from feeding point

$$= i\,r\left(l\,x - \frac{x^2}{2}\right)$$ [See Art. 13·4]

Here, $x = 150$ m

∴ Desired voltage drop $= 2 \times 0{\cdot}0006\left(200 \times 150 - \dfrac{150 \times 150}{2}\right) = \mathbf{22{\cdot}5\ V}$

(*ii*) Total current entering the distributor,

$$I = i \times l = 2 \times 200 = 400\text{ A}$$

Total resistance of the distributor,

$$R = r \times l = 0{\cdot}0006 \times 200 = 0{\cdot}12\,\Omega$$

∴ Total drop over the distributor

$$= \frac{1}{2}\,I\,R = \frac{1}{2} \times 400 \times 0{\cdot}12 = \mathbf{24\ V}$$

Example *Calculate the voltage at a distance of 200 m of a 300 m long distributor uniformly loaded at the rate of 0.75 A per metre. The distributor is fed at one end at 250 V. The resistance of the distributor (go and return) per metre is 0·00018 Ω. Also find the power loss in the distributor.*

Solution.

Voltage drop at a distance x from supply end

$$= i\,r\left(l\,x - \frac{x^2}{2}\right)$$

Here $i = 0{\cdot}75$ A/m; $l = 300$ m ; $x = 200$ m ; $r = 0{\cdot}00018\ \Omega$/m

$$\therefore \quad \text{Voltage drop} = 0{\cdot}75 \times 0{\cdot}00018\left[300 \times 200 - \frac{(200)^2}{2}\right] = 5.4\text{ V}$$

Voltage at a distance of 200 m from supply end

$$= 250 - 5{\cdot}4 = 244{\cdot}6\text{ V}$$

Power loss in the distributor is

$$P = \frac{i^2\,r\,l^3}{3} = \frac{(0{\cdot}75)^2 \times 0{\cdot}00018 \times (300)^3}{3} = 911{\cdot}25\text{ W}$$

TUTORIAL PROBLEMS

1. A 2-wire d.c. distributor 500 m long is loaded uniformly at the rate of 0·4A/m. If the voltage drop in the distributor is not to exceed 5V, calculate the area of X-section of each conductor required when the distributor is fed at one end. Take resistivity of conductor material as $1{\cdot}7 \times 10^{-8}\ \Omega$m. [3·4 cm²]
2. A uniformly distributed load on a distributor of length 500 m is rated at 1 A per metre length. The distributor is fed from one end at 220V. Determine the voltage drop at a distance of 400 m from the feeding point. Assume a loop resistance of $2 \times 10^{-5}\ \Omega$ per metre. [2·4 V]
3. A 250 m, 2-wire d.c. distributor fed from one end is loaded uniformly at the rate of 0·8 A per metre. The resistance of each conductor is 0·0002 Ω per metre. Find the necessary voltage at the feeding point to maintain 250 V at the far end of the distributor. [260 V]

Distributor Fed at Both Ends — Concentrated Loading

Whenever possible, it is desirable that a long distributor should be fed at both ends instead of at one end only, since total voltage drop can be considerably reduced without increasing the cross-section of the conductor. The two ends of the distributor may be supplied with (*i*) equal voltages (*ii*) unequal voltages.

(*i*) **Two ends fed with equal voltages.** Consider a distributor AB fed at both ends with equal voltages V volts and having concentrated loads I_1, I_2, I_3, I_4 and I_5 at points C, D, E, F and G respectively as shown in Fig. 13.14. As we move away from one of the feeding points, *say* A, p.d. goes on decreasing till it reaches the minimum value at some load point, *say* E, and then again starts rising and becomes V volts as we reach the other feeding point B.

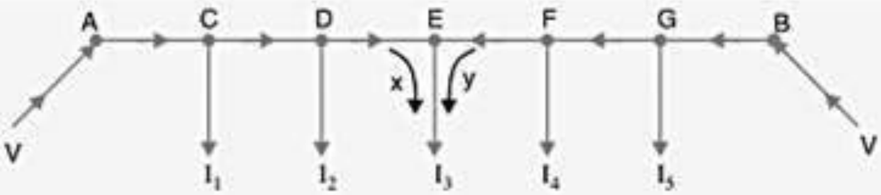


Fig. 13.14

All the currents tapped off between points A and E (minimum p.d. point) will be supplied from the feeding point A while those tapped off between B and E will be supplied from the feeding point B. The current tapped off at point E itself will be partly supplied from A and partly from B. If these currents are x and y respectively, then,

$$I_3 = x + y$$

Therefore, we arrive at a very important conclusion that at the point of minimum potential, current comes from both ends of the distributor.

Point of minimum potential. It is generally desired to locate the point of minimum potential. There is a simple method for it. Consider a distributor AB having three concentrated loads I_1, I_2 and I_3 at points C, D and E respectively. Suppose that current supplied by feeding end A is I_A. Then current distribution in the various sections of the distributor can be worked out as shown in Fig. 13.15 (*i*). Thus

$$I_{AC} = I_A; \qquad I_{CD} = I_A - I_1$$
$$I_{DE} = I_A - I_1 - I_2; \qquad I_{EB} = I_A - I_1 - I_2 - I_3$$

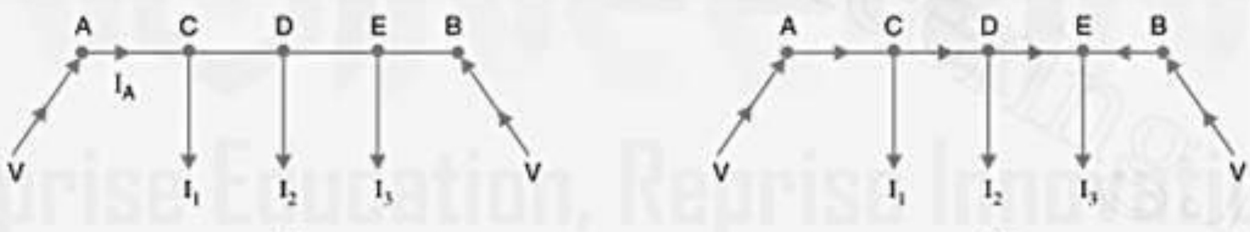


Fig. 13.15

Voltage drop between A and B = Voltage drop over AB

or $\quad V - V = I_A R_{AC} + (I_A - I_1) R_{CD} + (I_A - I_1 - I_2) R_{DE} + (I_A - I_1 - I_2 - I_3) R_{EB}$

From this equation, the unknown I_A can be calculated as the values of other quantities are generally given. Suppose *actual* directions of currents in the various sections of the distributor are indicated as shown in Fig. 13.15 (*ii*). The load point where the currents are coming from both sides of the distributor is the point of minimum potential *i.e.* point E in this case

(*ii*) **Two ends fed with unequal voltages.** Fig. 13.16 shows the distributor AB fed with unequal voltages ; end A being fed at V_1 volts and end B at V_2 volts. The point of minimum potential can be found by following the same procedure as discussed above. Thus in this case,

Voltage drop between A and B = Voltage drop over AB

or $\quad V_1 - V_2$ = Voltage drop over AB

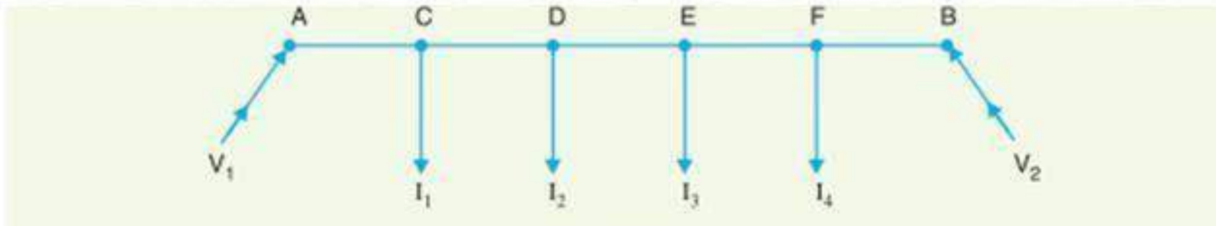


Fig. 13.16

Example *A 2-wire d.c. street mains AB, 600 m long is fed from both ends at 220 V. Loads of 20 A, 40 A, 50 A and 30 A are tapped at distances of 100m, 250m, 400m and 500 m from the end A respectively. If the area of X-section of distributor conductor is 1cm², find the minimum consumer voltage. Take* $\rho = 1{\cdot}7 \times 10^{-6}\ \Omega\ cm$.

Solution. Fig. 13.17 shows the distributor with its tapped currents. Let I_A amperes be the current supplied from the feeding end A. Then currents in the various sections of the distributor are as shown in Fig. 13.17.

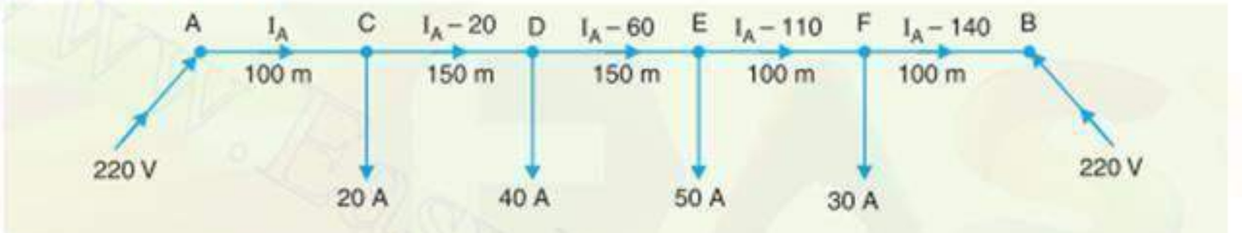


Fig. 13.17

Resistance of 1 m length of distributor

$$= 2 \times \frac{1{\cdot}7 \times 10^{-6} \times 100}{1} = 3{\cdot}4 \times 10^{-4}\ \Omega$$

$$\begin{aligned}
\text{Resistance of section } AC, R_{AC} &= (3{\cdot}4 \times 10^{-4}) \times 100 = 0{\cdot}034\ \Omega \\
\text{Resistance of section } CD, R_{CD} &= (3{\cdot}4 \times 10^{-4}) \times 150 = 0{\cdot}051\ \Omega \\
\text{Resistance of section } DE, R_{DE} &= (3{\cdot}4 \times 10^{-4}) \times 150 = 0{\cdot}051\ \Omega \\
\text{Resistance of section } EF, R_{EF} &= (3{\cdot}4 \times 10^{-4}) \times 100 = 0{\cdot}034\ \Omega \\
\text{Resistance of section } FB, R_{FB} &= (3{\cdot}4 \times 10^{-4}) \times 100 = 0{\cdot}034\ \Omega
\end{aligned}$$

$$\begin{aligned}
\text{Voltage at } B &= \text{Voltage at } A - \text{Drop over length } AB \\
\text{or} \qquad V_B &= V_A - [I_A R_{AC} + (I_A - 20) R_{CD} + (I_A - 60) R_{DE} \\
&\qquad + (I_A - 110) R_{EF} + (I_A - 140) R_{FB}] \\
\text{or} \qquad 220 &= 220 - [0{\cdot}034 I_A + 0{\cdot}051 (I_A - 20) + 0{\cdot}051 (I_A - 60) \\
&\qquad + 0{\cdot}034 (I_A - 110) + 0{\cdot}034 (I_A - 140)] \\
&= 220 - [0{\cdot}204 I_A - 12{\cdot}58] \\
\text{or} \qquad 0{\cdot}204 I_A &= 12{\cdot}58 \\
\therefore \qquad I_A &= 12{\cdot}58/0{\cdot}204 = 61{\cdot}7\ \text{A}
\end{aligned}$$

The *actual distribution of currents in the various sections of the distributor is shown in Fig. 13.18. It is clear that currents are coming to load point E from both sides *i.e.* from point D and point F. Hence, E is the point of minimum potential.

$\therefore$ Minimum consumer voltage,

$$V_E = V_A - [I_{AC} R_{AC} + I_{CD} R_{CD} + I_{DE} R_{DE}]$$

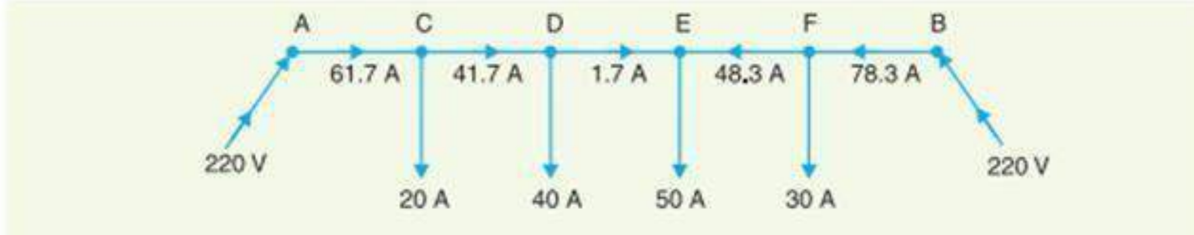


Fig. 13.18

$$= 220 - [61{\cdot}7 \times 0{\cdot}034 + 41{\cdot}7 \times 0{\cdot}051 + 1{\cdot}7 \times 0{\cdot}051]$$
$$= 220 - 4{\cdot}31 = \mathbf{215{\cdot}69\ V}$$

Example *A 2-wire d.c. distributor AB is fed from both ends. At feeding point A, the voltage is maintained as at 230 V and at B 235 V. The total length of the distributor is 200 metres and loads are tapped off as under :*

25 A at 50 metres from A ; 50 A at 75 metres from A

30 A at 100 metres from A ; 40 A at 150 metres from A

The resistance per kilometre of one conductor is 0·3 Ω. Calculate :

(*i*) *currents in various sections of the distributor*

(*ii*) *minimum voltage and the point at which it occurs*

Solution. Fig. 13.19 shows the distributor with its tapped currents. Let I_A amperes be the current supplied from the feeding point A. Then currents in the various sections of the distributor are as shown in Fig 13.19.

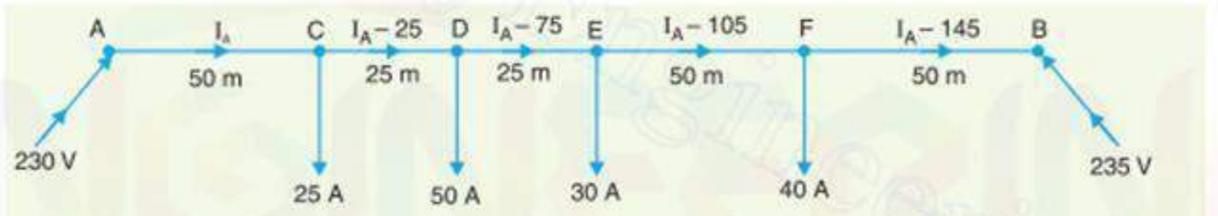


Fig. 13.19

Resistance of 1000 m length of distributor (both wires)

$$= 2 \times 0{\cdot}3 = 0{\cdot}6\,\Omega$$

Resistance of section AC, $R_{AC} = 0{\cdot}6 \times 50/1000 = 0{\cdot}03\,\Omega$

Resistance of section CD, $R_{CD} = 0{\cdot}6 \times 25/1000 = 0{\cdot}015\,\Omega$

Resistance of section DE, $R_{DE} = 0{\cdot}6 \times 25/1000 = 0{\cdot}015\,\Omega$

Resistance of section EF, $R_{EF} = 0{\cdot}6 \times 50/1000 = 0{\cdot}03\,\Omega$

Resistance of section FB, $R_{FB} = 0{\cdot}6 \times 50/1000 = 0{\cdot}03\,\Omega$

Voltage at B = Voltage at A − Drop over AB

or
$$V_B = V_A - [I_A R_{AC} + (I_A - 25) R_{CD} + (I_A - 75) R_{DE} + (I_A - 105) R_{EF} + (I_A - 145) R_{FB}]$$

or
$$235 = 230 - [0{\cdot}03\,I_A + 0{\cdot}015\,(I_A - 25) + 0{\cdot}015\,(I_A - 75) + 0{\cdot}03\,(I_A - 105) + 0{\cdot}03\,(I_A - 145)]$$

or
$$235 = 230 - [0{\cdot}12\,I_A - 9]$$

∴
$$I_A = \frac{239 - 235}{0{\cdot}12} = 33{\cdot}34\text{ A}$$

(*i*) ∴ Current in section AC, $I_{AC} = I_A = \mathbf{33{\cdot}34\ A}$

Current in section CD, $I_{CD} = I_A - 25 = 33{\cdot}34 - 25 = \mathbf{8{\cdot}34\ A}$

$$\text{Current in section } DE, I_{DE} = I_A - 75 = 33{\cdot}34 - 75 = -41{\cdot}66 \text{ A } \textit{from D to E}$$
$$= \mathbf{41{\cdot}66\ A} \textit{ from E to D}$$
$$\text{Current in section } EF, I_{EF} = I_A - 105 = 33{\cdot}34 - 105 = -71{\cdot}66 \text{ A } \textit{from E to F}$$
$$= \mathbf{71{\cdot}66\ A} \textit{ from F to E}$$
$$\text{Current in section } FB, I_{FB} = I_A - 145 = 33{\cdot}34 - 145 = -111{\cdot}66 \text{ A } \textit{from F to B}$$
$$= \mathbf{111{\cdot}66\ A} \textit{ from B to F}$$

(*ii*) The actual distribution of currents in the various sections of the distributor is shown in Fig. 13.20. The currents are coming to load point D from both sides of the distributor. Therefore, load point D is the point of minimum potential.

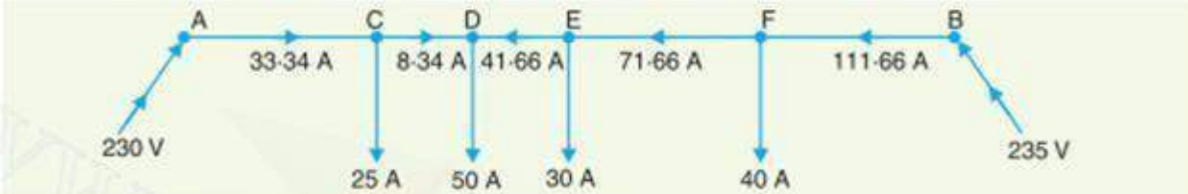


Fig. 13.20

$$\text{Voltage at } D, V_D = V_A - [I_{AC} R_{AC} + I_{CD} R_{CD}]$$
$$= 230 - [33{\cdot}34 \times 0{\cdot}03 + 8{\cdot}34 \times 0{\cdot}015]$$
$$= 230 - 1{\cdot}125 = \mathbf{228{\cdot}875\ V}$$

Example *A two-wire d.c. distributor AB, 600 metres long is loaded as under :*

Distance from A (metres) :	*150*	*300*	*350*	*450*
Loads in Amperes :	*100*	*200*	*250*	*300*

The feeding point A is maintained at 440 V and that of B at 430 V. If each conductor has a resistance of 0·01 Ω per 100 metres, calculate :

(*i*) *the currents supplied from A to B,* (*ii*) *the power dissipated in the distributor.*

Solution. Fig. 13.21 shows the distributor with its tapped currents. Let I_A amperes be the current supplied from the feeding point A. Then currents in the various sections of the distributor are as shown in Fig.13.21.

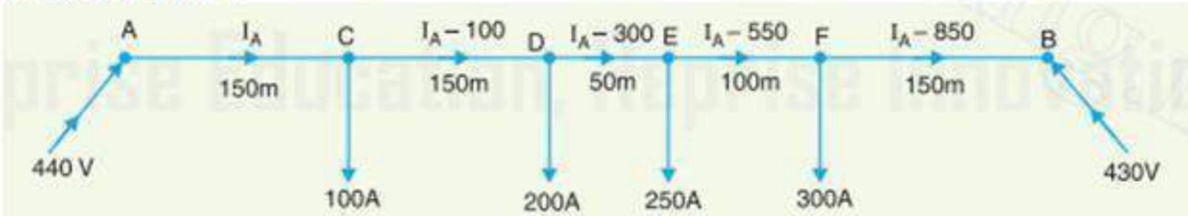


Fig. 13.21

Resistance of 100 m length of distributor (both wires)

$$= 2 \times 0{\cdot}01 = 0{\cdot}02\,\Omega$$

$$\text{Resistance of section } AC, R_{AC} = 0{\cdot}02 \times 150/100 = 0{\cdot}03\,\Omega$$
$$\text{Resistance of section } CD, R_{CD} = 0{\cdot}02 \times 150/100 = 0{\cdot}03\,\Omega$$
$$\text{Resistance of section } DE, R_{DE} = 0{\cdot}02 \times 50/100 = 0{\cdot}01\,\Omega$$
$$\text{Resistance of section } EF, R_{EF} = 0{\cdot}02 \times 100/100 = 0{\cdot}02\,\Omega$$
$$\text{Resistance of section } FB, R_{FB} = 0{\cdot}02 \times 150/100 = 0{\cdot}03\,\Omega$$

$$\text{Voltage at } B = \text{Voltage at } A \text{ — Drop over } AB$$

or

$$V_B = V_A - [I_A R_{AC} + (I_A - 100) R_{CD} + (I_A - 300) R_{DE} + (I_A - 550) R_{EF} + (I_A - 850) R_{FB}]$$

or $$430 = 440 - [0{\cdot}03\,I_A + 0{\cdot}03\,(I_A - 100) + 0{\cdot}01\,(I_A - 300) + 0{\cdot}02\,(I_A - 550) + 0{\cdot}03\,(I_A - 850)]$$

or $$430 = 440 - [0{\cdot}12\,I_A - 42{\cdot}5]$$

$\therefore$ $$I_A = \frac{482\cdot5 - 430}{0\cdot12} = 437{\cdot}5\text{ A}$$

The actual distribution of currents in the various sections of the distributor is shown in Fig.13.22. Incidentally, E is the point of minimum potential.

(*i*) Referring to Fig. 13.22, it is clear that

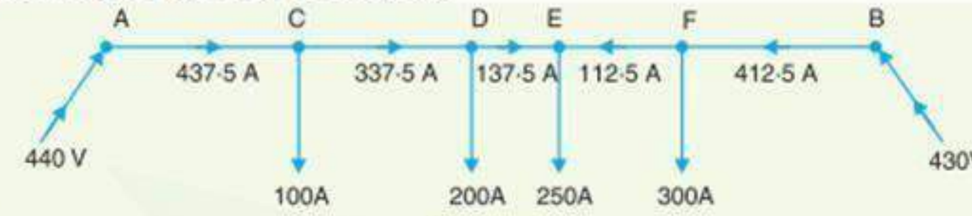


Fig. 13.22

Current supplied from end A, I_A = **437·5 A**

Current supplied from end B, I_B = **412·5 A**

(*ii*) Power loss in the distributor

$= I_{AC}^2 R_{AC} + I_{CD}^2 R_{CD} + I_{DE}^2 R_{DE} + I_{EF}^2 R_{EF} + I^2_{FB} R_{FB}$

$= (437{\cdot}5)^2 \times 0{\cdot}03 + (337{\cdot}5)^2 \times 0{\cdot}03 + (137{\cdot}5)^2 \times 0{\cdot}01 + (112{\cdot}5)^2 \times 0{\cdot}02 + (412{\cdot}5)^2 \times 0{\cdot}03$

$= 5742 + 3417 + 189 + 253 + 5104 = 14{,}705$ watts = **14·705 kW**

Example *An electric train runs between two sub-stations 6 km apart maintained at voltages 600 V and 590 V respectively and draws a constant current of 300 A while in motion. The track resistance of go and return path is 0·04 Ω/km. Calculate :*

(*i*) *the point along the track where minimum potential occurs*

(*ii*) *the current supplied by the two sub-stations when the train is at the point of minimum potential*

Solution. The single line diagram is shown in Fig. 13.23 where substation A is at 600 V and sub-station B at 590 V. Suppose that minimum potential occurs at point M at a distance x km from the substation A. Let I_A amperes be the current supplied by the sub-station A. Then current supplied by sub-station B is $300 - I_A$ as shown in Fig 13.23.

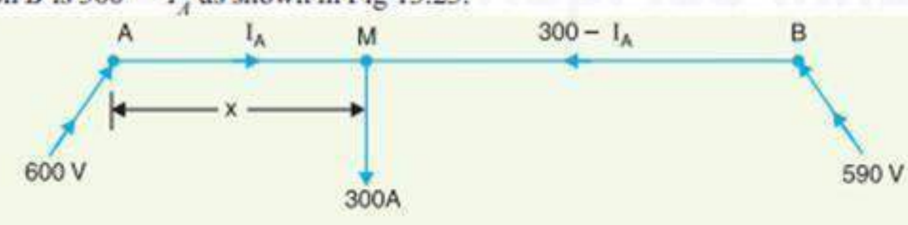


Fig. 13.23

Resistance of track (go and return path) per km

$= 0{\cdot}04\,\Omega$

Track resistance for section AM, $R_{AM} = 0{\cdot}04x\ \Omega$

Track resistance for section MB, $R_{MB} = 0{\cdot}04\,(6-x)\Omega$

Potential at M, $V_M = V_A - I_A R_{AM}$ ... (*i*)

Also, Potential at M, $V_M = V_B - (300 - I_A)\,R_{MB}$ ... (*ii*)

From equations (*i*) and (*ii*), we get,

$$V_A - I_A R_{AM} = V_B - (300 - I_A) R_{MB}$$

or $$600 - 0{\cdot}04x\, I_A = 590 - (300 - I_A) \times 0{\cdot}04\,(6 - x)$$

or $$600 - 0{\cdot}04x\, I_A = 590 - 0{\cdot}04\,(1800 - 300\,x - 6\,I_A + I_A \times x)$$

or $$600 - 0{\cdot}04x\, I_A = 590 - 72 + 12\,x + 0{\cdot}24\,I_A - 0{\cdot}04xI_A$$

or $$0{\cdot}24 I_A = 82 - 12\,x$$

or $$I_A = 341{\cdot}7 - 50\,x$$

Substituting the value of I_A in eq. (*i*), we get,

$$V_M = V_A - (341{\cdot}7 - 50\,x) \times 0{\cdot}04x$$

$\therefore$ $$V_M = 600 - 13{\cdot}7x + 2x^2 \qquad ...(iii)$$

(*i*) For V_M to be minimum, its differential coefficient *w.r.t.* x must be zero *i.e.*

$$\frac{d}{dx}(600 - 13{\cdot}7x + 2x^2) = 0$$

or $$0 - 13{\cdot}7 + 4x = 0$$

$\therefore$ $$x = 13{\cdot}7/4 = \mathbf{3{\cdot}425\ km}$$

i.e. minimum potential occurs at a distance of 3·425 km from the sub-station *A*.

(*ii*) $\therefore$ Current supplied by sub-station *A*

$$= 341{\cdot}7 - 50 \times 3{\cdot}425 = 341{\cdot}7 - 171{\cdot}25 = \mathbf{170{\cdot}45\ A}$$

Current supplied by sub-station *B* $= 300 - I_A = 300 - 170{\cdot}45 = \mathbf{129{\cdot}55\ A}$

TUTORIAL PROBLEMS

1. A 2-wire d.c. distributor *A B* is fed at both ends at the same voltage of 230 V. The length of the distributor is 500 metres and the loads are tapped off from the end A as shown below :

Load :	100 A	60 A	40 A	100 A
Distance :	50 m	150 m	250 m	400 m

If the maximum voltage drop of 5·5 V is to be allowed, find the X-sectional area of each conductor and point of minimum potential. Specific resistance of conductor material may be taken as $1{\cdot}73 \times 10^{-8}\ \Omega$ m. **[1·06 cm^2 ; 250 m from A]**

2. A d.c. distributor *A B* is fed at both ends. At feeding point *A*, the voltage is maintained at 235 V and at *B* at 236 V. The total length of the distributor is 200 metres and loads are tapped off as under :

20 A at 50 m from *A*
40 A at 75 m from *A*
25 A at 100 m from *A*
30 A at 150 m from *A*

The resistance per kilometre of one conductor is 0·4 Ω. Calculate the minimum voltage and the point at which it occurs. **[232·175 V ; 75 m from point A]**

3. A two conductor main *A B*, 500 m in length is fed from both ends at 250 volts. Loads of 50 A, 60 A, 40 A and 30 A are tapped at distance of 100 m, 250 m, 350 m and 400 m from end *A* respectively. If the X-section of conductor be 1 cm^2 and specific resistance of the material of the conductor is 1·7 μ Ω cm, determine the minimum consumer voltage. **[245·07 V]**

Ring Distributor

A distributor arranged to form a closed loop and fed at one or more points is called a *ring distributor*. Such a distributor starts from one point, makes a loop through the area to be served, and returns to the original point. For the purpose of calculating voltage distribution, the distributor can be considered as consisting of a series of open distributors fed at both ends. The principal advantage of ring distributor is that by proper choice in the number of feeding points, great economy in copper can be affected.

The most simple case of a ring distributor is the one having only one feeding point as shown in Fig. 13.36(*ii*). Here A is the feeding point and tappings are taken from points B and C. For the purpose of calculations, it is equivalent to a straight distributor fed at both ends with equal voltages.

Example *A 2-wire d.c. ring distributor is 300 m long and is fed at 240 V at point A. At point B, 150 m from A, a load of 120 A is taken and at C, 100 m in the opposite direction, a load of 80 A is taken. If the resistance per 100 m of single conductor is 0·03 Ω, find :*

(*i*) *current in each section of distributor*

(*ii*) *voltage at points B and C*

Solution.

Resistance per 100 m of distributor

$$= 2 \times 0{\cdot}03 = 0{\cdot}06\,\Omega$$

Resistance of section AB, $R_{AB} = 0{\cdot}06 \times 150/100 = 0{\cdot}09\,\Omega$

Resistance of section BC, $R_{BC} = 0{\cdot}06 \times 50/100 = 0{\cdot}03\,\Omega$

Resistance of section CA, $R_{CA} = 0{\cdot}06 \times 100/100 = 0{\cdot}06\,\Omega$

(*i*) Let us suppose that a current I_A flows in section AB of the distributor. Then currents in sections BC and CA will be $(I_A - 120)$ and $(I_A - 200)$ respectively as shown in Fig. 13.36 (*i*).

According to Kirchhoff's voltage law, the voltage drop in the closed loop $ABCA$ is zero *i.e.*

$$I_{AB} R_{AB} + I_{BC} R_{BC} + I_{CA} R_{CA} = 0$$

or $$0{\cdot}09\,I_A + 0{\cdot}03\,(I_A - 120) + 0{\cdot}06\,(I_A - 200) = 0$$

or $$0{\cdot}18\,I_A = 15{\cdot}6$$

$\therefore$ $$I_A = 15{\cdot}6/0{\cdot}18 = 86{\cdot}67 \text{ A}$$

The actual distribution of currents is as shown in Fig. 13.36 (*ii*) from where it is seen that B is the point of minimum potential.

Fig. 13.36

Current in section AB, $I_{AB} = I_A =$ **86·67 A** *from A to B*

Current in section BC, $I_{BC} = I_A - 120 = 86{\cdot}67 - 120 = -\,33{\cdot}33$ A

$= $ **33.33 A** *from C to B*

Current in section CA, $I_{CA} = I_A - 200 = 86{\cdot}67 - 200 = -\,113{\cdot}33$ A

$= $ **113 ·33 A** *from A to C*

(*ii*) Voltage at point B, $V_B = V_A - I_{AB} R_{AB} = 240 - 86{\cdot}67 \times 0{\cdot}09 =$ **232·2 V**

Voltage at point C, $V_C = V_B + I_{BC} R_{BC}$

$= 232 \cdot 2 + 33 \cdot 33 \times 0 \cdot 03 =$ **233·2 V**

Example *A 2-wire d.c. distributor ABCDEA in the form of a ring main is fed at point A at 220 V and is loaded as under :*

10A at B ; 20A at C ; 30A at D and 10 A at E.

The resistances of various sections (go and return) are : $AB = 0 \cdot 1\ \Omega$; $BC = 0 \cdot 05\ \Omega$; $CD = 0 \cdot 01\ \Omega$; $DE = 0 \cdot 025\ \Omega$ and $EA = 0 \cdot 075\ \Omega$. Determine :

(*i*) *the point of minimum potential*

(*ii*) *current in each section of distributor*

Solution. Fig. 13.37 (*i*) shows the ring main distributor. Let us suppose that current I flows in section AB of the distributor. Then currents in the various sections of the distributor are as shown in Fig. 13.37 (*i*).

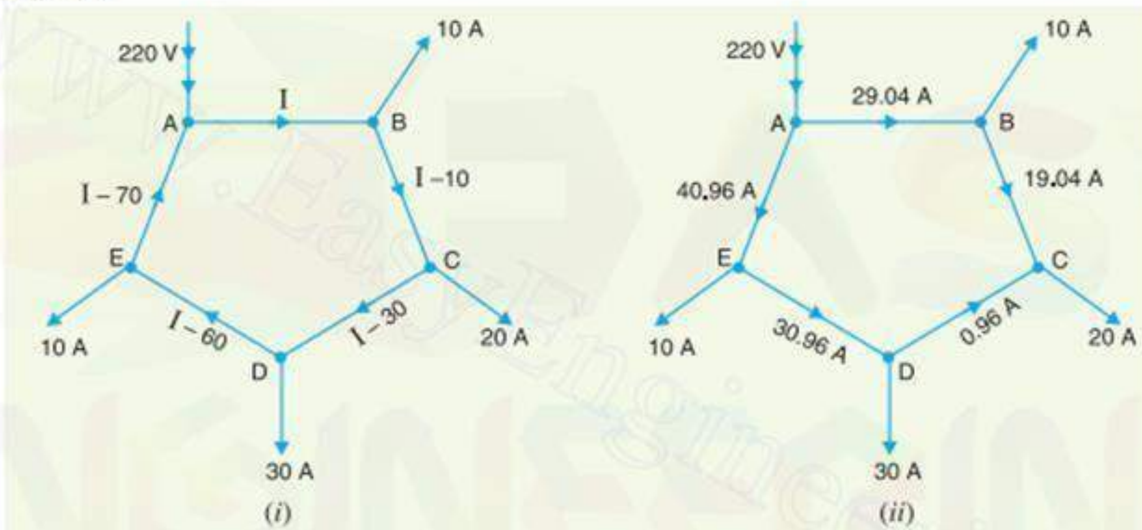


Fig. 13.37

(*i*) According to Kirchhoff's voltage law, the voltage drop in the closed loop $ABCDEA$ is zero *i.e.*

$$I_{AB} R_{AB} + I_{BC} R_{BC} + I_{CD} R_{CD} + I_{DE} R_{DE} + I_{EA} R_{EA} = 0$$

or $\quad 0 \cdot 1I + 0 \cdot 05\,(I - 10) + 0 \cdot 01\,(I - 30) + 0 \cdot 025\,(I - 60) + 0 \cdot 075\,(I - 70) = 0$

or $\quad 0 \cdot 26\,I = 7 \cdot 55$

$\therefore \quad I = 7 \cdot 55 / 0 \cdot 26 = 29 \cdot 04$ A

The actual distribution of currents is as shown in Fig. 13.37 (*ii*) from where it is clear that C is the point of minimum potential.

$\therefore$ C is the point of minimum potential.

(*ii*) Current in section AB $= I =$ **29·04 A** *from A to B*

Current in section BC $= I - 10 = 29 \cdot 04 - 10 =$ **19·04 A** *from B to C*

Current in section CD $= I - 30 = 29 \cdot 04 - 30 = -\,0 \cdot 96$ A $=$ **0·96 A** *from D to C*

Current in section DE $= I - 60 = 29 \cdot 04 - 60 = -\,30 \cdot 96$ A $=$ **30·96 A** *from E to D*

Current in section EA $= I - 70 = 29 \cdot 04 - 70 = -\,40 \cdot 96$ A $=$ **40·96 A** *from A to E*

A.C. Distribution Calculations

A.C. distribution calculations differ from those of d.c. distribution in the following respects :

(*i*) In case of d.c. system, the voltage drop is due to resistance alone. However, in a.c. system, the voltage drops are due to the combined effects of resistance, inductance and capacitance.

(*ii*) In a d.c. system, additions and subtractions of currents or voltages are done arithmetically but in case of a.c. system, these operations are done vectorially.

(*iii*) In an a.c. system, power factor (p.f.) has to be taken into account. Loads tapped off form the distributor are generally at different power factors. There are two ways of referring power factor *viz*

(*a*) It may be referred to supply or receiving end voltage which is regarded as the reference vector.

(*b*) It may be referred to the voltage at the load point itself.

There are several ways of solving a.c. distribution problems. However, symbolic notation method has been found to be most convenient for this purpose. In this method, voltages, currents and impedances are expressed in complex notation and the calculations are made exactly as in d.c. distribution.

Methods of Solving A.C. Distribution Problems

In a.c. distribution calculations, power factors of various load currents have to be considered since currents in different sections of the distributor will be the vector sum of load currents and not the arithmetic sum. The power factors of load currents may be given (*i*) *w.r.t.* receiving or sending end voltage or (*ii*) *w.r.t.* to load voltage itself. Each case shall be discussed separately.

(*i*) **Power factors referred to receiving end voltage.** Consider an a.c. distributor AB with concentrated loads of I_1 and I_2 tapped off at points C and B as shown in Fig. 14.1. Taking the receiving end voltage V_B as the reference vector, let lagging power factors at C and B be $\cos\phi_1$ and $\cos\phi_2$ *w.r.t.* V_B. Let R_1, X_1 and R_2, X_2 be the resistance and reactance of sections AC and CB of the distributor.

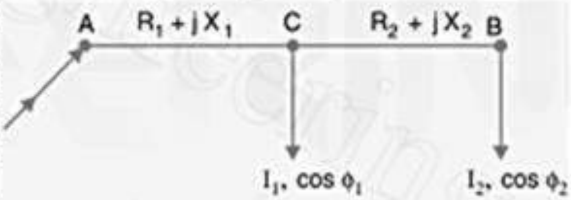


Fig. 14.1

Impedance of section AC, $\overline{Z_{AC}} = R_1 + j X_1$

Impedance of section CB, $\overline{Z_{CB}} = R_2 + j X_2$

Load current at point C, $\vec{I_1} = I_1 (\cos\phi_1 - j \sin\phi_1)$

Load current at point B, $\vec{I_2} = I_2 (\cos\phi_2 - j \sin\phi_2)$

Current in section CB, $\overline{I_{CB}} = \vec{I_2} = I_2 (\cos\phi_2 - j \sin\phi_2)$

Current in section AC, $\overline{I_{AC}} = \vec{I_1} + \vec{I_2}$

$$= I_1 (\cos\phi_1 - j \sin\phi_1) + I_2 (\cos\phi_2 - j \sin\phi_2)$$

Voltage drop in section CB, $\overline{V_{CB}} = \overline{I_{CB}}\,\overline{Z_{CB}} = I_2 (\cos\phi_2 - j \sin\phi_2)(R_2 + j X_2)$

Voltage drop in section AC, $\overline{V_{AC}} = \overline{I_{AC}}\,\overline{Z_{AC}} = \left(\vec{I_1} + \vec{I_2}\right) Z_{AC}$

$$= [I_1(\cos\phi_1 - j\sin\phi_1) + I_2(\cos\phi_2 - j\sin\phi_2)]\,[R_1 + jX_1]$$

Sending end voltage, $\qquad \overrightarrow{V_A} = \overrightarrow{V_B} + \overrightarrow{V_{CB}} + \overrightarrow{V_{AC}}$

Sending end current, $\qquad \overrightarrow{I_A} = \overrightarrow{I_1} + \overrightarrow{I_2}$

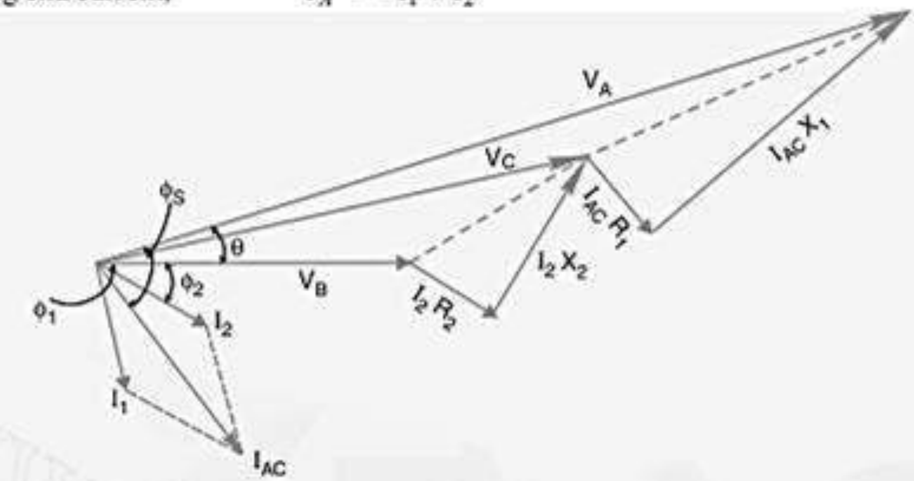


Fig. 14.2

The vector diagram of the a.c. distributor under these conditions is shown in Fig. 14.2. Here, the receiving end voltage V_B is taken as the reference vector. As power factors of loads are given *w.r.t.* V_B, therefore, I_1 and I_2 lag behind V_B by ϕ_1 and ϕ_2 respectively.

(*ii*) Power factors referred to respective load voltages. Suppose the power factors of loads in the previous Fig. 14.1 are referred to their respective load voltages. Then ϕ_1 is the phase angle between V_C and I_1 and ϕ_2 is the phase angle between V_B and I_2. The vector diagram under these conditions is shown in Fig. 14.3.

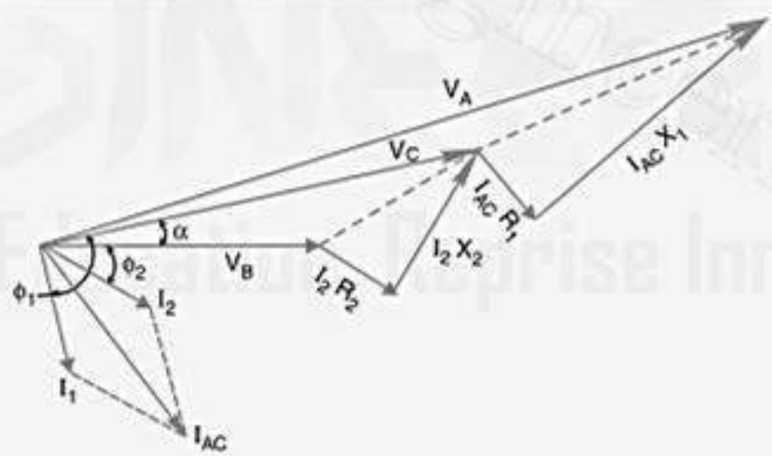


Fig. 14.3

Voltage drop in section CB $= \overrightarrow{I_2}\,\overrightarrow{Z_{CB}} = I_2(\cos\phi_2 - j\sin\phi_2)(R_2 + jX_2)$

Voltage at point C $= \overrightarrow{V_B}$ + Drop in section $CB = V_C \angle \alpha$ (say)

Now $\qquad \overrightarrow{I_1} = I_1 \angle -\phi_1 \quad$ *w.r.t.* voltage V_C

$\therefore \qquad \overrightarrow{I_1} = I_1 \angle -(\phi_1 - \alpha) \qquad$ *w.r.t.* voltage V_B

i.e. $\qquad \overrightarrow{I_1} = I_1[\cos(\phi_1 - \alpha) - j\sin(\phi_1 - \alpha)]$

Now $\qquad \overrightarrow{I_{AC}} = \overrightarrow{I_1} + \overrightarrow{I_2}$

$$= I_1 [\cos(\phi_1 - \alpha) - j \sin(\phi_1 - \alpha)] + I_2 (\cos\phi_2 - j \sin\phi_2)$$

Voltage drop in section $AC = \overrightarrow{I_{AC}}\ \overrightarrow{Z_{AC}}$

$\therefore$ Voltage at point $A = V_B$ + Drop in CB + Drop in AC

Example *A single phase a.c. distributor AB 300 metres long is fed from end A and is loaded as under :*

(i) 100 A at 0·707 p.f. lagging 200 m from point A

(ii) 200 A at 0·8 p.f. lagging 300 m from point A

The load resistance and reactance of the distributor is 0·2 Ω and 0·1 Ω per kilometre. Calculate the total voltage drop in the distributor. The load power factors refer to the voltage at the far end.

Solution. Fig. 14.4 shows the single line diagram of the distributor.

Impedance of distributor/km $= (0 \cdot 2 + j\,0 \cdot 1)\,\Omega$

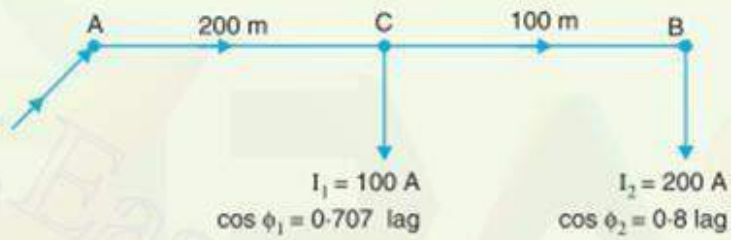


Fig. 14.4

Impedance of section AC, $\overrightarrow{Z_{AC}} = (0 \cdot 2 + j\,0 \cdot 1) \times 200/1000 = (0 \cdot 04 + j\,0 \cdot 02)\,\Omega$

Impedance of section CB, $\overrightarrow{Z_{CB}} = (0 \cdot 2 + j\,0 \cdot 1) \times 100/1000 = (0 \cdot 02 + j\,0 \cdot 01)\,\Omega$

Taking voltage at the far end B as the reference vector, we have,

Load current at point B, $\overrightarrow{I_2} = I_2 (\cos\phi_2 - j \sin\phi_2) = 200\,(0 \cdot 8 - j\,0 \cdot 6)$
$= (160 - j\,120)$ A

Load current at point C, $\overrightarrow{I_1} = I_1 (\cos\phi_1 - j \sin\phi_1) = 100\,(0 \cdot 707 - j\,0 \cdot 707)$
$= (70 \cdot 7 - j\,70 \cdot 7)$ A

Current in section CB, $\overrightarrow{I_{CB}} = \overrightarrow{I_2} = (160 - j\,120)$ A

Current in section AC, $\overrightarrow{I_{AC}} = \overrightarrow{I_1} + \overrightarrow{I_2} = (70 \cdot 7 - j\,70 \cdot 7) + (160 - j\,120)$
$= (230 \cdot 7 - j\,190 \cdot 7)$ A

Voltage drop in section CB, $\overrightarrow{V_{CB}} = \overrightarrow{I_{CB}}\ \overrightarrow{Z_{CB}} = (160 - j\,120)(0 \cdot 02 + j\,0 \cdot 01)$
$= (4 \cdot 4 - j\,0 \cdot 8)$ volts

Voltage drop in section AC, $\overrightarrow{V_{AC}} = \overrightarrow{I_{AC}}\ \overrightarrow{Z_{AC}} = (230 \cdot 7 - j\,190 \cdot 7)(0 \cdot 04 + j\,0 \cdot 02)$
$= (13 \cdot 04 - j\,3 \cdot 01)$ volts

Voltage drop in the distributor $= \overrightarrow{V_{AC}} + \overrightarrow{V_{CB}} = (13 \cdot 04 - j\,3 \cdot 01) + (4 \cdot 4 - j\,0 \cdot 8)$
$= (17 \cdot 44 - j\,3 \cdot 81)$ volts

Magnitude of drop $= \sqrt{(17 \cdot 44)^2 + (3 \cdot 81)^2} =$ **17·85 V**

Example *A single phase distributor 2 kilometres long supplies a load of 120 A at 0·8 p.f. lagging at its far end and a load of 80 A at 0·9 p.f. lagging at its mid-point. Both power factors are* *referred to the voltage at the far end. The resistance and reactance per km (go and return) are 0·05 Ω and 0·1 Ω respectively. If the voltage at the far end is maintained at 230 V, calculate :*

(*i*) *voltage at the sending end*

(*ii*) *phase angle between voltages at the two ends.*

Solution. Fig. 14.5 shows the distributor *AB* with *C* as the mid-point

Impedance of distributor/km $= (0{\cdot}05 + j\,0{\cdot}1)\,\Omega$

Impedance of section *AC*, $\vec{Z}_{AC} = (0{\cdot}05 + j\,0{\cdot}1) \times 1000/1000 = (0{\cdot}05 + j\,0{\cdot}1)\,\Omega$

Impedance of section *CB*, $\vec{Z}_{CB} = (0{\cdot}05 + j\,0{\cdot}1) \times 1000/1000 = (0{\cdot}05 + j\,0{\cdot}1)\,\Omega$

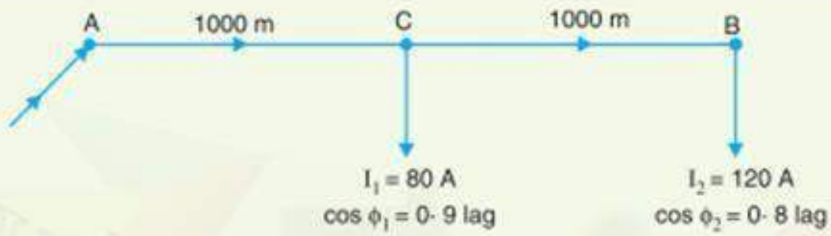


Fig. 14.5

Let the voltage V_B at point *B* be taken as the reference vector.

Then, $\overrightarrow{V_B} = 230 + j\,0$

(*i*) Load current at point *B*, $\overrightarrow{I_2} = 120\,(0{\cdot}8 - j\,0{\cdot}6) = 96 - j\,72$

Load current at point *C*, $\overrightarrow{I_1} = 80\,(0{\cdot}9 - j\,0{\cdot}436) = 72 - j\,34{\cdot}88$

Current in section *CB*, $\overrightarrow{I_{CB}} = \overrightarrow{I_2} = 96 - j\,72$

Current in section *AC*, $\overrightarrow{I_{AC}} = \overrightarrow{I_1} + \overrightarrow{I_2} = (72 - j\,34{\cdot}88) + (96 - j\,72)$

$= 168 - j\,106{\cdot}88$

Drop in section *CB*, $\overrightarrow{V_{CB}} = \overrightarrow{I_{CB}}\;\overrightarrow{Z_{CB}} = (96 - j\,72)\,(0{\cdot}05 + j\,0{\cdot}1)$

$= 12 + j\,6$

Drop in section *AC*, $\overrightarrow{V_{AC}} = \overrightarrow{I_{AC}}\;\overrightarrow{Z_{AC}} = (168 - j\,106{\cdot}88)\,(0{\cdot}05 + j\,0{\cdot}1)$

$= 19{\cdot}08 + j\,11{\cdot}45$

∴ Sending end voltage, $\overrightarrow{V_A} = \overrightarrow{V_B} + \overrightarrow{V_{CB}} + \overrightarrow{V_{AC}}$

$= (230 + j\,0) + (12 + j\,6) + (19.08 + j\,11.45)$

$= 261.08 + j\,17.45$

Its magnitude is $= \sqrt{(261{\cdot}08)^2 + (17{\cdot}45)^2} =$ **261·67 V**

(*ii*) The phase difference θ between V_A and V_B is given by :

$$\tan\theta = \frac{17{\cdot}45}{261{\cdot}08} = 0{\cdot}0668$$

∴ $\theta = \tan^{-1} 0{\cdot}0668 =$ **3·82°**

Example *A single phase distributor one km long has resistance and reactance per conductor of 0·1 Ω and 0·15 Ω respectively. At the far end, the voltage V_B = 200 V and the current is 100 A at a p.f. of 0·8 lagging. At the mid-point M of the distributor, a current of 100 A is tapped at a p.f.*

of 0·6 lagging with reference to the voltage V_M at the mid-point. Calculate :

(*i*) *voltage at mid-point*

(*ii*) *sending end voltage V_A*

(*iii*) *phase angle between V_A and V_B*

Solution. Fig. 14.6 shows the single line diagram of the distributor *AB* with *M* as the mid-point.

Total impedance of distributor $= 2(0{\cdot}1 + j\,0{\cdot}15) = (0{\cdot}2 + j\,0{\cdot}3)\,\Omega$

Impedance of section *AM*, $\overrightarrow{Z_{AM}} = (0{\cdot}1 + j\,0{\cdot}15)\,\Omega$

Impedance of section *MB*, $\overrightarrow{Z_{MB}} = (0{\cdot}1 + j\,0{\cdot}15)\,\Omega$

Let the voltage V_B at point *B* be taken as the reference vector.

Then, $\overrightarrow{V_B} = 200 + j\,0$

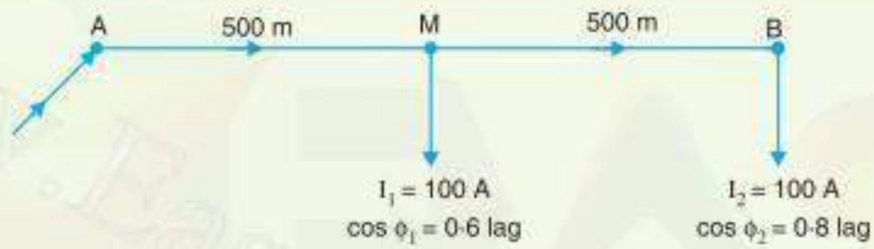


Fig. 14.6

(*i*) Load current at point *B*, $\overrightarrow{I_2} = 100\,(0{\cdot}8 - j\,0{\cdot}6) = 80 - j\,60$

Current in section *MB*, $\overrightarrow{I_{MB}} = \overrightarrow{I_2} = 80 - j\,60$

Drop in section *MB*,
$$\overrightarrow{V_{MB}} = \overrightarrow{I_{MB}}\,\overrightarrow{Z_{MB}}$$
$$= (80 - j\,60)\,(0{\cdot}1 + j\,0{\cdot}15) = 17 + j\,6$$

$\therefore$ Voltage at point *M*,
$$\overrightarrow{V_M} = \overrightarrow{V_B} + \overrightarrow{V_{MB}} = (200 + j\,0) + (17 + j\,6)$$
$$= 217 + j\,6$$

Its magnitude is $= \sqrt{(217)^2 + (6)^2} =$ **217·1 V**

Phase angle between V_M and V_B, $\alpha = \tan^{-1} 6/217 = \tan^{-1} 0{\cdot}0276 = 1{\cdot}58^\circ$

(*ii*) The load current I_1 has a lagging p.f. of 0·6 *w.r.t.* V_M. It lags behind V_M by an angle $\phi_1 = \cos^{-1} 0{\cdot}6 = 53{\cdot}13^\circ$

$\therefore$ Phase angle between I_1 and V_B, $\phi'_1 = \phi_1 - \alpha = 53{\cdot}13^\circ - 1{\cdot}58 = 51{\cdot}55^\circ$

Load current at *M*,
$$\overrightarrow{I_1} = I_1\,(\cos\phi_1' - j\sin\phi_1') = 100\,(\cos 51{\cdot}55^\circ - j\sin 51{\cdot}55^\circ)$$
$$= 62{\cdot}2 - j\,78{\cdot}3$$

Current in section *AM*,
$$\overrightarrow{I_{AM}} = \overrightarrow{I_1} + \overrightarrow{I_2} = (62{\cdot}2 - j\,78{\cdot}3) + (80 - j\,60)$$
$$= 142{\cdot}2 - j\,138{\cdot}3$$

Drop in section *AM*,
$$\overrightarrow{V_{AM}} = \overrightarrow{I_{AM}}\,\overrightarrow{Z_{AM}} = (142{\cdot}2 - j\,138{\cdot}3)\,(0{\cdot}1 + j\,0{\cdot}15)$$
$$= 34{\cdot}96 + j\,7{\cdot}5$$

Sending end voltage,
$$\overrightarrow{V_A} = \overrightarrow{V_M} + \overrightarrow{V_{AM}} = (217 + j\,6) + (34{\cdot}96 + j\,7{\cdot}5)$$

$$= 251{\cdot}96 + j\,13{\cdot}5$$

Its magnitude is $= \sqrt{(251{\cdot}96)^2 + (13{\cdot}5)^2} = \mathbf{252{\cdot}32\ V}$

(*iii*) The phase difference θ between V_A and V_B is given by :

$$\tan\theta = 13{\cdot}5/251{\cdot}96 = 0{\cdot}05358$$

$$\therefore \quad \theta = \tan^{-1} 0{\cdot}05358 = \mathbf{3{\cdot}07^\circ}$$

Hence supply voltage is 252·32 V and leads V_B by 3·07°.

Four-Wire Star-Connected Unbalanced Loads

We can obtain this type of load in two ways. First, we may connect a 3-phase, 4-wire unbalanced load to a 3-phase, 4-wire supply as shown in Fig. 14.10. Note that star point *N* of the supply is connected to the load star point *N′*. Secondly, we may connect single phase loads between any line and the neutral wire as shown in Fig.14.11. This will also result in a 3-phase, 4-wire **unbalanced load because it is rarely possible that single phase loads on all the three phases have the same magnitude and power factor. Since the load is unbalanced, the line currents will be different in magnitude and displaced from one another by unequal angles. The current in the neutral wire will be the phasor sum of the three line currents *i.e.*

Current in neutral wire, $\quad I_N = I_R + I_Y + I_B \quad$ *...phasor sum*

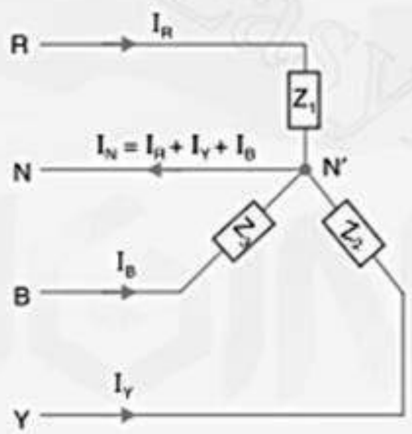


Fig. 14.10

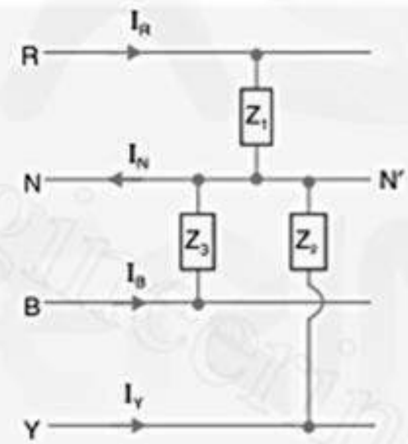


Fig. 14.11

The following points may be noted carefully :

(*i*) Since the neutral wire has negligible resistance, supply neutral *N* and load neutral *N′* will be at the same potential. It means that voltage across each impedance is equal to the phase voltage of the supply. However, current in each phase (or line) will be different due to unequal impedances.

(*ii*) The amount of current flowing in the neutral wire will depend upon the magnitudes of line currents and their phasor relations. In most circuits encountered in practice, the neutral current is equal to or smaller than one of the line currents. The exceptions are those circuits having severe unbalance.

Chapter7 Underground cables

An **underground cable** *essentially consists of one or more conductors covered with suitable insulation and surrounded by a protecting cover.*

Although several types of cables are available, the type of cable to be used will depend upon the working voltage and service requirements. In general, a cable must fulfil the following necessary requirements:

- The conductor used in cables should be tinned stranded copper or aluminium of high conductivity. Stranding is done so that conductor may become flexible and carry more current.
- The conductor size should be such that the cable carries the desired load current without overheating and causes voltage drop within permissible limits.
- The cable must have proper thickness of insulation in order to give high degree of safety and reliability at the voltage for which it is designed.
- The cable must be provided with suitable mechanical protection so that it may withstand the rough use in laying it.
- The materials used in the manufacture of cables should be such that there is complete chemical and physical stabilitythroughout.

Construction of cables:

- ***Cores or Conductors.*** A cable may have one or more than one core (conductor) depending upon the type of service for which it is intended. For instance, the 3-conductor cable shown in Fig. 11.1 is used for 3-phase service. The conductors are made of tinned copper or aluminium and are usually stranded in order to provide flexibility to the cable.
- ***Insulation.*** Each core or conductor is provided with a suitable thickness of insulation, the thickness of layer depending upon the voltage to be withstood by the cable. The commonly used materials for insulation are impregnated paper, varnished cambric or rubber mineral compound.
- **Metallic sheath**: In order to protect the cable from moisture, gases and other damaging liquids (acids or alkalis) in the soiland atmosphere, a metallic sheath of lead or aluminium is provided over the insulation
- **Bedding:** Over the metallic sheath is applied a layer of bedding which consists of a fibrous material like jute or hessian tape. The purpose of bedding is to protect the metallic sheath against corrosion and from mechanical injury due to armouring.
- ***Armouring***. Over the bedding, armouring is provided which consists of one or two layers of galvanised steel wire or steel tape. Its purpose is to protect the cable from mechanical injury while laying it and during the course of handling. Armouring may not be done in the case of some cables
- ***Serving.*** In order to protect armouring from atmospheric conditions, a layer of fibrous material (like jute) similar to bedding is provided over the armouring. This is known as *serving.*

Insulating materials for cables:

the insulating materials used in cables should have the following properties :

- High insulation resistance to avoid leakagecurrent.
- High dielectric strength to avoid electrical breakdown of thecable.
- High mechanical strength to withstand the mechanical handling of cables.
- Non-hygroscopic *i.e.*, it should not absorb moisture from air or soil. The moisture tends to decrease the insulation resistance and hastens the breakdown of the cable. In case the insulating material is hygroscopic, it must be enclosed in a waterproof covering like lead sheath.
- Non-inflammable.

- Low cost so as to make the underground system a viable proposition.
- unaffected by acids and alkalies to avoid any chemical action

The principal insulating materials used in cables are rubber, vulcanised India rubber, impregnated paper, varnished cambric and polyvinyl chloride.

Rubber. Rubber may be obtained from milky sap of tropical trees or it may be produced from oil products. It has relative permittivity varying between 2 and 3, dielectric strength is about 30 kV/mm and resistivity of insulation is $10^{17}\Omega$ cm. Although pure rubber has reasonably high insulating properties, it suffers from some major drawbacks *viz.,* readily absorbs moisture, maximum safe temperature is low (about 38ºC), soft and liable to damage due to rough handling and ages when exposed to light. Therefore, pure rubber cannot be used as an insulating material.

Vulcanised India Rubber (*V.I.R.*). It is prepared by mixing pure rubber with mineral mat- ter such as zine oxide, red lead etc., and 3 to 5% of sulphur. The compound so formed is rolled into thin sheets and cut into strips. The rubber compound is then applied to the conductor and is heated to a temperature of about 150ºC. The whole process is called *vulcanisation* and the product obtained is known as vulcanised India rubber.

Vulcanised India rubber has greater mechanical strength, durability and wear resistant property than pure rubber. Its main drawback is that sulphur reacts very quickly with copper and for this reason, cables using *VIR* insulation have tinned copper conductor. The *VIR* insulation is generally used for low and moderate voltage cables.

Impregnated paper. It consists of chemically pulped paper made from wood chippings and impregnated with some compound such as paraffinic or napthenic material. This type of insulation has almost superseded the rubber insulation. It is because it has the advantages of low cost, low capacitance, high dielectric strength and high insulation resistance. The only disadvantage is that paper is hygroscopic and even if it is impregnated with suitable compound, it absorbs moisture and thus lowers the insulation resistance of the cable. For this reason, paper insulated cables are always provided with some protective covering and are never left unsealed. If it is required to be left unused on the site during laying, its ends are temporarily covered with wax or tar.

Varnished cambric. It is a cotton cloth impregnated and coated with varnish. This type of insulation is also known as *empire tape.* The cambric is lapped on to the conductor in the form of a tape and its surfaces are coated with petroleum jelly compound to allow for the sliding of one turn over another as the cable is bent. As the varnished cambric is hygroscopic, therefore, such cables are always provided with metallic sheath. Its dielectric strength is about 4 kV/mm and permittivity is 2.5 to 3.8.

Polyvinyl chloride (*PVC*). This insulating material is a synthetic compound. It is obtained from the polymerisation of acetylene and is in the form of white powder. For obtaining this material as a cable insulation, it is compounded with certain materials known as plasticizers which are liquids with high boiling point. The plasticizer forms a gell and renders the material plastic over the desired range of temperature.

Classification of cables:

Cables for underground service may be classified in two ways according to (*i*) the type of insulating material used in their manufacture (*ii*) the voltage for which they are manufactured. However, the latter method of classification is generally preferred, according to which cables can be divided into the following groups :

- Low-tension (L.T.) cables — upto 1000 V

- High-tension (H.T.) cables — upto 11,000V
- Super-tension (S.T.) cables — from 22 kV to 33 kV
- Extra high-tension (E.H.T.) cables — from 33 kV to 66 kV
- Extra super voltage cables — beyond 132 kV

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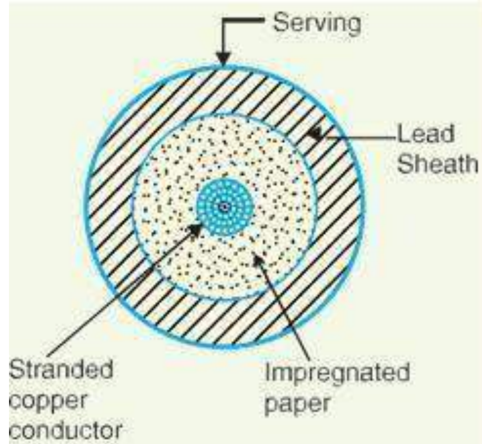


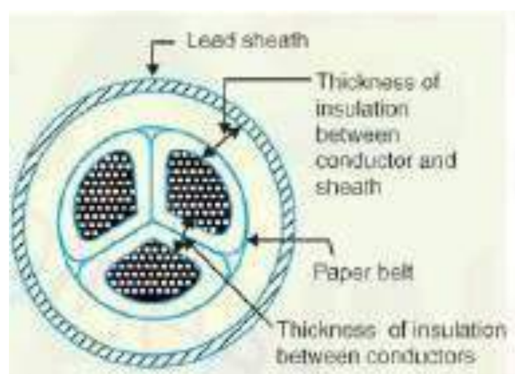
A cable may have one or more than one core depending upon the type of service for which it is intended. It may be (*i*) single-core (*ii*) two-core (*iii*) three-core (*iv*) four-core etc. For a 3-phase service, either 3-single-core cables or three-core cable can be used depending upon the operating voltage and load demand. The cable has ordinary construction be- cause the stresses developed in the cable for low voltages (upto 6600 V) are generally small. It consists of one circular core of tinned stranded copper (or aluminium) insulated by layers of impregnated paper. The insulation is surrounded by a lead sheath which prevents the entry of moisture into the inner parts. In order to protect the lead sheath from corrosion, an overall serving of compounded fibrous material (jute etc.) is provided. Single-core cables are not usually armoured in order to avoid excessive sheath losses. The principal advantages of single-core cables are simple construction and availability of larger copper section.

Cables for 3 phase service:

underground cables are generally required to deliver 3-phase power. For the purpose, either three-core cable or three single core cables may be used. For voltages upto 66 kV, 3-core cable (*i.e.*, multi-core construction) is preferred due to economic reasons. However, for voltages beyond 66 kV, 3-core-cables become too large and unwieldy and, therefore, single-core cables are used. The following types of cables are generally used for 3-phase service :

1. Belted cables — upto 11 kV
2. Screened cables — from 22 kV to 66 kV
3. Pressure cables — beyond 66 kV.

Belted cables. These cables are used for voltages upto 11kV but in extraordinary cases, their use may be extended upto 22kV. The cores are insulated from each other by layers of impregnated paper. Another layer of impregnated paper tape, called *paper belt* is wound round the grouped insulated cores. The gap between the insulated cores is filled with fibrous insulating material (jute etc.) so as to give circular cross-section to the cable. The cores are generally stranded and may be of non- circular shape to make better use of available space. The belt is covered with lead sheath to protect the cable against ingress of moisture and mechanical injury. The lead sheath is covered with one or more layers of armouring with an outer serving.



Screened cables. These cables are meant for use upto 33 kV, but in particular cases their use may be extended to operating voltages upto 66 kV. Two principal types of screened cables are H- type cables and S.L. type cables.

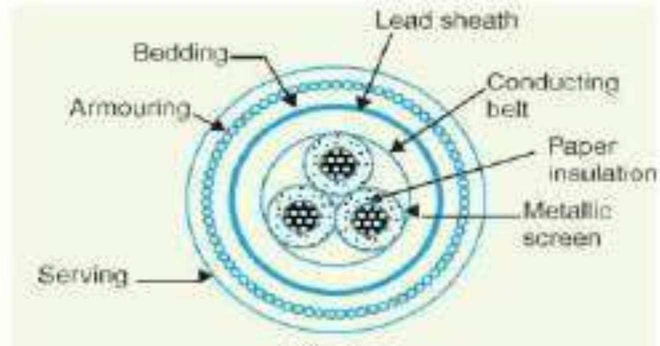


Fig. 11.4

Pressure cables For voltages beyond 66 kV, solid type cables are unreliable because there is a danger of breakdown of insulation due to the presence of voids. When the operating voltages are greater than 66 kV, *pressure cables* are used. In such cables, voids are eliminated by increasing the pressure of compound and for this reason they are called pressure cables. Two types of pressure cables *viz* oil-filled cables and gas pressure cables are commonly used.

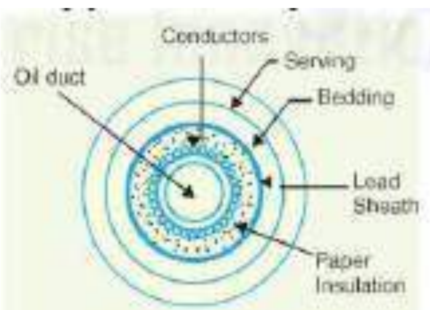


Fig. 11.6 Single-core conductor channel, oil-filled cable

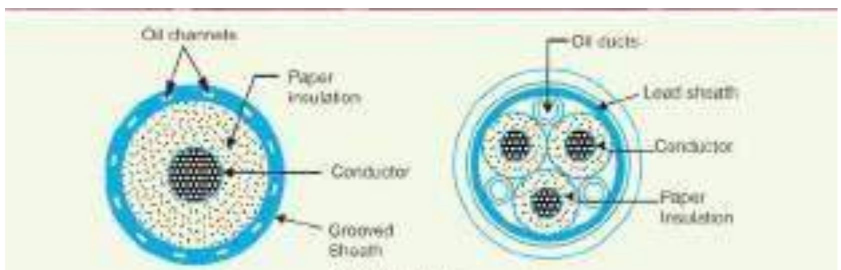


Fig. 11.7 *Oil-filled cables* **Fig. 11.8**

Laying of underground cable:

The reliability of underground cable network depends to a considerable extent upon the proper laying and attachment of fittings *i.e.*, cable end boxes, joints, branch connectors etc. There are three main methods of laying underground cables *viz.*, direct laying, draw-in system and the solid system.

Direct laying. This method of laying underground cables is simple and cheap and is much favoured in modernpractice. In this method, a trench of about 1·5 metres deep and 45 cm wide is dug. The trench is covered with a layer of fine sand (of about 10 cm thickness) and the cable is laid over this sand bed. The sand prevents the entry of moisture from the ground and thus protects the cable from decay. After the cable has been laid in the trench, it is covered with another layer of sand of about 10 cm thickness.

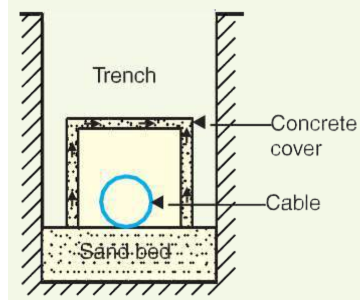


Fig. 11.10

The trench is then covered with bricks and other materials in order to protect the cable from mechanical injury. When more than one cable is to be laid in the same trench, a horizontal or vertical inter- axial spacing of at least 30 cm is provided in order to reduce the effect of mutual heating and also to ensure that a fault occurring on one cable does not damage the adjacent cable. Cables to be laid in this way must have serving of bituminised paper and hessian tape so as to provide protection against corrosion and electorlysis.

Advantages

- It is a simple and less costlymethod.
- It gives the best conditions for dissipating the heat generated in the cables.
- It is a clean and safe method as the cable is invisible and free from external disturbances.

Disadvantages

- The extension of load is possible only by a completely new excavation which may cost as much as the original work.
- The alterations in the cable netwok cannot be madeeasily.
- The maintenance cost is veryhigh.
- Localisation of fault is difficult.
- It cannot be used in congested areas where excavation is expensive and inconvenient.

Draw-insystem. In this method, conduit or duct of glazed stone or cast iron or concrete are laid in the ground with manholes at suitable positions along

the cable route. The cables are then pulled into position from manholes. Fig. 11.11 shows section through four-way underground duct line. Three of the ducts carry transmission cables and the fourth duct carries relay protection connection, pilot wires. Care must be taken that where the duct line changes direction ; depths, dips and offsets be made with a very long radius or it will be difficult to pull a large cable between the manholes. The distance between the manholes should not be too long so as to simplify thepulling in of the cables. The cables to be laid in this way need not be armoured but must be provided with serving of hessian and jute in order to protect them when being pulled into the ducts.

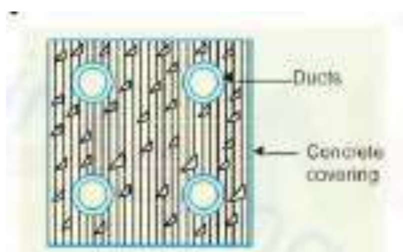


Fig. 11.11

Advantages

- Repairs, alterations or additions to the cable network can be made without opening the ground.
- As the cables are not armoured, therefore, joints become simpler and maintenance cost is reduced considerably.
- There are very less chances of fault occurrence due to strong mechanical protection pro- vided by the system.

Disadvantages

- The initial cost is very high.
- The current carrying capacity of the cables is reduced due to the close grouping of cables and unfavourable conditions for dissipation of heat.

Solid system. In this method of laying, the cable is laid in open pipes or troughs dug out in earth along the cable route. The troughing is of cast iron, stoneware, asphalt or treated wood. After the cable is laid in position, the troughing is filled with a bituminous or asphaltic compound and covered over. Cables laid in this manner are usually plain lead covered because troughing affords good mechanical protection.

Disadvantages

- It is more expensive than direct laid system.
- It requires skilled labour and favourable weather conditions.
- Due to poor heat dissipation facilities, the current carrying capacity of the cable is reduced

Types of cable faults:

Cables are generally laid directly in the ground or in ducts in the underground distribution system. For this reason, there are little chances of faults in underground cables. However, if a fault does occur, it is difficult to locate and repair the fault because conductors are not visible. Nevertheless, the following are the faults most likely to occur in underground cables :

(i) Open-circuit fault
(ii) Short-circuit fault
(iii) Earth fault.

Open-circuit fault. When there is a break in the conductor of a cable, it is called open- circuit fault. The open-circuit fault can be checked by a megger. For this purpose, the three conductors of the 3-core cable at the far end are shorted and earthed. Then resistance be- tween each conductor and earth is measured by a megger. The megger will indicate zero resistance in the circuit of the conductor that is not broken. However, if the conductor is broken, the megger will indicate infinite resistance in its circuit.

Short-circuit fault. When two conductors of a multi-core cable come in electrical contact with each other due to insulation failure, it is called a short-circuit fault. Again, we can seek the help of a megger to check this fault. For this purpose, the two terminals of themegger are connected to any two conductors. If the megger gives zero reading, it indicates short- circuit fault between these conductors. The same step is repeated for other conductors taking two at a time.

Earth fault. When the conductor of a cable comes in contact with earth, it is called earth fault or ground fault. To identify this fault, one terminal of the megger is connected to the conductor and the other terminal connected to earth. If the megger indicates zero reading, it means the conductor is earthed. The same procedure is repeated for other conductors of the cable.

Loop tests for location of faults in underground cables:

(i) Murray loop test
(ii) Varley loop test

Murray Loo p Test

The Murray loop test is the most common and accurate method of locating earth fault or short-circuit fault in underground cables.

(i) **Earth fault :** Fig. 11.22 shows the circuit diagram for locating the earth fault by Murray loop test. Here *AB* is the sound cable and *CD* is the faulty cable; the earth fault occuring at point *F*. The far end *D* of the faulty cable is joined to the far end *B* of the sound cable through a low resistance link. Two variable resistances P and Q are joined to ends A and C (See Fig. 11.22) respectively and serve as the ratio arms of the Wheatstone bridge.

Let R = resistance of the conductor loop upto the fault from the test end

X = resistance of the other length of the loop

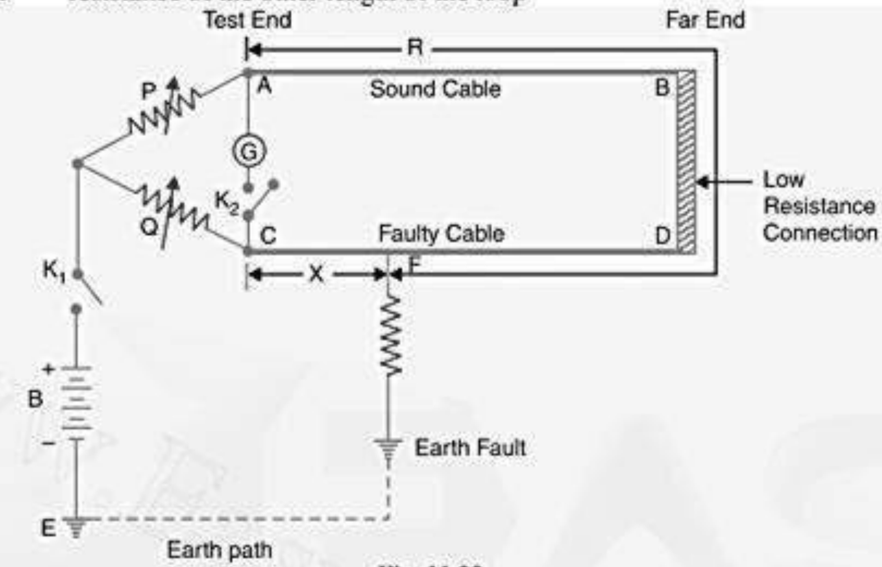


Fig. 11.22

Note that P, Q, R and X are the four arms of the Wheatstone bridge. The resistances P and Q are varied till the galvanometer indicates zero deflection.

In the balanced position of the bridge, we have,

$$\frac{P}{Q} = \frac{R}{X}$$

or $$\frac{P}{Q}+1 = \frac{R}{X}+1$$

or $$\frac{P+Q}{Q} = \frac{R+X}{X}$$

If r is the resistance of each cable, then $R + X = 2r$.

$\therefore$ $$\frac{P+Q}{Q} = \frac{2r}{X}$$

or $$X = \frac{Q}{P+Q}\times 2r$$

If l is the length of each cable in metres, then resistance per metre length of cable = $\frac{r}{l}$.

$\therefore$ Distance of fault point from test end is

$$d = \frac{X}{r/l} = \frac{Q}{P+Q}\times 2r\times\frac{l}{r} = \frac{Q}{P+Q}\times 2l$$

or $$d = \frac{Q}{P+Q}\times \text{(loop length) *metres}$$

Thus the position of the fault is located. Note that resistance of the fault is in the battery circuit and not in the bridge circuit. Therefore, fault resistance does not affect the balancing of the bridge. However, if the fault resistance is high, the sensitivity of the bridge is reduced.

(*ii*) **Short-circuit fault :** Fig. 11.23 shows the circuit diagram for locating the short-circuit fault by Murray loop test. Again P, Q, R and X are the four arms of the bridge. Note that fault resistance is in the battery circuit and not in the bridge circuit. The bridge in balanced by adjusting the resistances P and Q. In the balanced position of the bridge :

$$\frac{P}{Q} = \frac{R}{X}$$

or
$$\frac{P+Q}{Q} = \frac{R+X}{X} = \frac{2r}{X}$$

$\therefore$
$$X = \frac{Q}{P+Q} \times 2r$$

or
$$X = \frac{Q}{P+Q} \times \text{(loop length) metres}$$

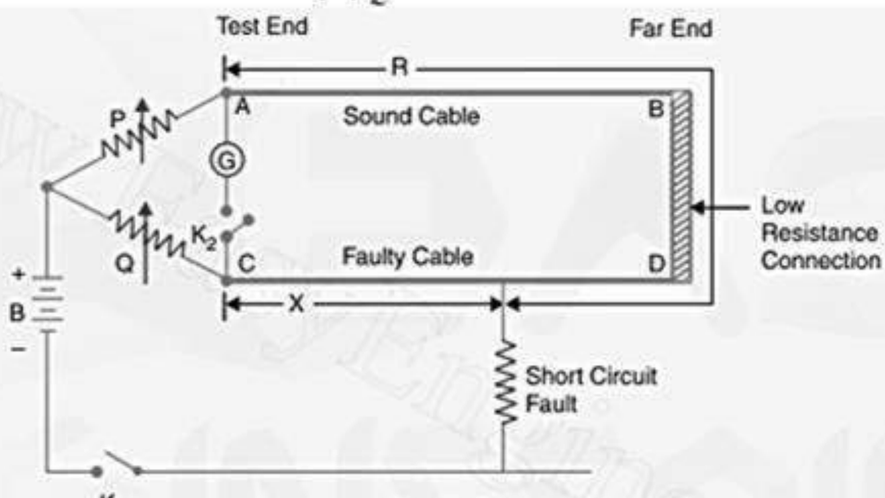


Fig. 11.23

Thus the position of the fault is located.

Varley Loop Test

The Varley loop test is also used to locate earth fault or short-circuit fault in underground cables. This test also employs Wheatstone bridge principle. It differs from Murray loop test in that here the ratio arms P and Q are fixed resistances. Balance is obtained by adjusting the variable resistance S

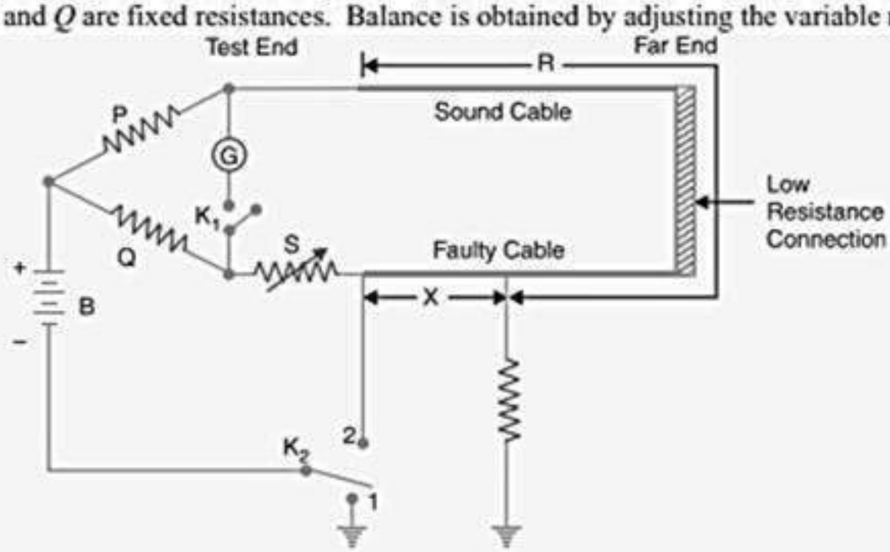


Fig. 11.24 Varley Loop Test (Earth Fault)

connected to the test end of the faulty cable. The connection diagrams for locating the earth fault and short-circuit fault by Varley loop test are shown in Figs. 11.24 and 11.25 respectively.

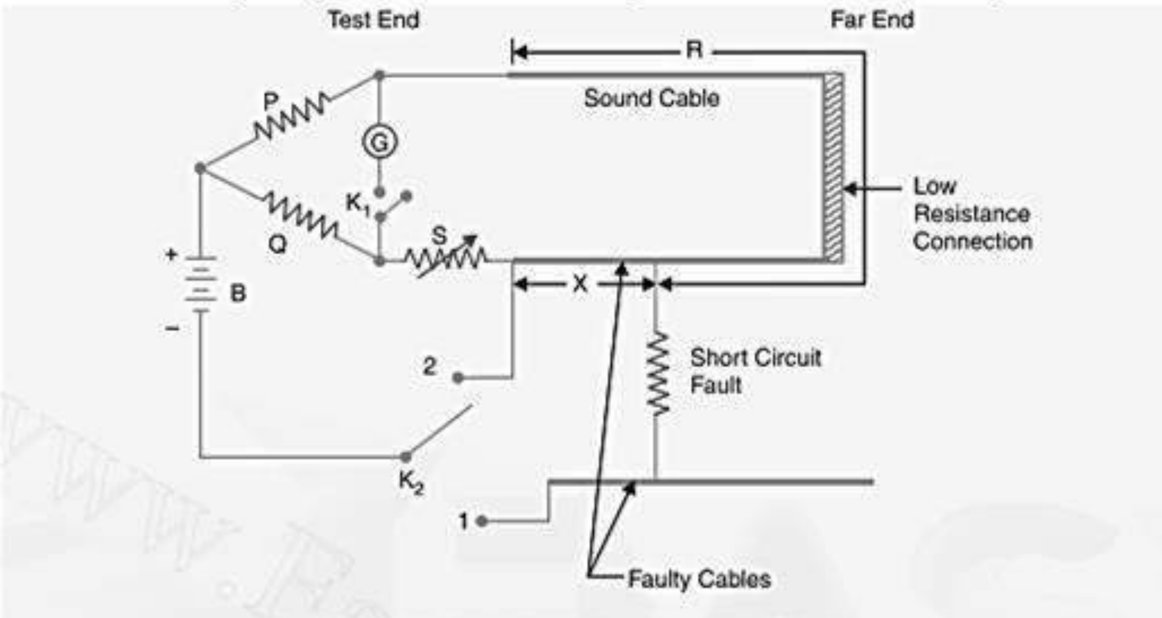


Fig. 11.25 Varley Loop Test (Short Circuit Test)

For earth fault or short-circuit fault, the key K_2 is first thrown to position 1. The variable resistance S is varied till the bridge is balanced for resistance value of S_1. Then,

$$\frac{P}{Q} = \frac{R}{X + S_1}$$

or
$$\frac{P+Q}{Q} = \frac{R + X + S_1}{X + S_1}$$

or
$$X = \frac{Q(R + X) - PS_1}{P + Q} \qquad ...(i)$$

Now key K_2 is thrown to position 2 (for earth fault or short-circuit fault) and bridge is balanced with new value of resistance S_2. Then,

$$\frac{P}{Q} = \frac{R+X}{S_2}$$

or
$$(R + X)\,Q = PS_2 \qquad ...(ii)$$

From eqs. (*i*) and (*ii*), we get,

$$X = \frac{P\,(S_2 - S_1)}{P + Q}$$

Since the values of P, Q, S_1 and S_2 are known, the value of X can be determined.

$$\text{Loop resistance} = R + X = \frac{P}{Q}S_2$$

If r is the resistance of the cable per metre length, then,

Distance of fault from the test end is

$$d = \frac{X}{r} \text{ metres}$$

<u>**Chapter8 Economic Aspects**</u>

The cosine of angle between voltage and current in an a.c. circuit is known as **power factor.**

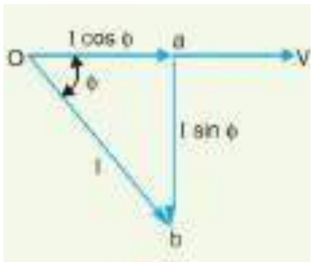


Fig. 6.1

Power factor= R/Z

Power factor= Active power / apparent power

Disadvantages of low power factor:

Lower the power factor, higher is the load current and *vice-versa*. A power factor less than unity results in the following disadvantages :

1. kVA rating of the equipment is inversely proportional to power factor. The smaller the power factor, the larger is the kVA rating. Therefore, at low power factor, the kVA rating of the equipment has to be made more, making the equipment larger and expensive.
2. the conductor will have to carry more current at low power factor. This necessitates large conductor size.
3. Large copper losses. The large current at low power factor causes more I^2R losses in all the elements of the supply system. This results in poor efficiency.
4. Poor voltage regulation. The large current at low lagging power factor causes greater voltage drops in alternators, transformers, transmission lines and distributors. This results in the decreased voltage available at the supply end, thus impairing the performance of utilisation devices.
5. Reduced handling capacity of system. The lagging power factor reduces the handling capacity of all the elements of the system. It is because the reactive component of current prevents the full utilisation of installed capacity.

Causes of low power factor:

(i) Most of the a.c. motors are of induction type (1ϕ and 3ϕ induction motors) which have low lagging power factor. These motors work at a power factor which is extremely small on light load (0·2 to 0·3) and rises to 0·8 or 0·9 at full load.

(ii) Arc lamps, electric discharge lamps and industrial heating furnaces operate at low lagging power factor.

(iii) The load on the power system is varying ; being high during morning and evening and low at other times. During low load period, supply voltage is increased which increases the magnetisation current. This results in the decreased power factor.

Power factor improvement:

The low power factor is mainly due to the fact that most of the power loads are inductive and, therefore, take lagging currents. In order to improve the power factor, some device taking leading power should be connected in parallel with the load. One of such devices can be a capacitor. The capacitor draws a leading current and partly or completely neutralises the lagging reactive component of load current. This raises the power factor of the load.

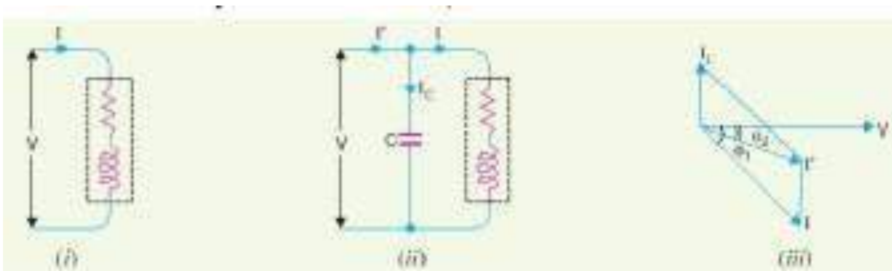


Fig. 6.3

Power Factor Improvement Equipment:

Normally, the power factor of the whole load on a large generating station is in the region of 0·8 to 0·9. However, sometimes it is lower and in such cases it is generally desirable to take special steps to improve the power factor. This can be achieved by the following equipment :

1. Static capacitors.**2.** Synchronous condenser.**3.** Phase advancers.

Static capacitor. The power factor can be improved by connecting capacitors in parallel with the equipment operating at lagging power factor. The capacitor (generally known as static capacitor) draws a leading current and partly or completely neutralises the lagging reactive component of load current. This raises the power factor of the load. For three-phase loads, the capacitors can be connected in delta or star.

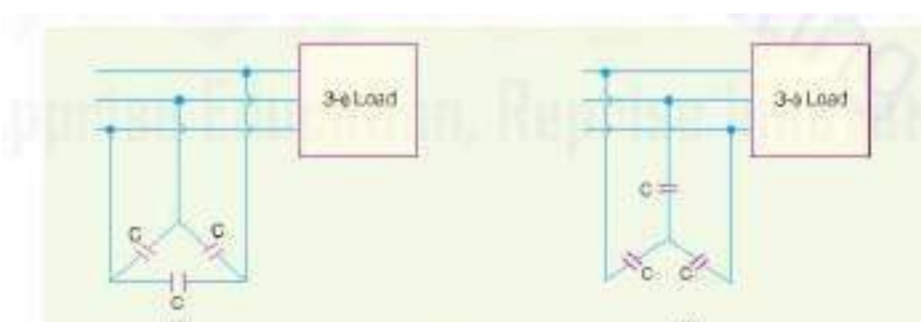


Fig. 6.4

Advantages

- They have low losses.
- They require little maintenance as there are no rotatingparts.
- They can be easily installed as they are light and require no foundation.

- They can work under ordinary atmospheric conditions.

Disadvantages

- They have short service life ranging from 8 to 10 years.
- They are easily damaged if the voltage exceeds the rated value.
- Once the capacitors are damaged, their repair is uneconomical.

Synchronous condenser. A synchronous motor takes a leading current when over-excited and, therefore, behaves as a capacitor. An over-excited synchronous motor running on no load is known as *synchronous condenser.* When such a machine is connected in parallel with thesupply, it takes a leading current which partly neutralises the lagging reactive component of the load. Thus the power factor is improved.

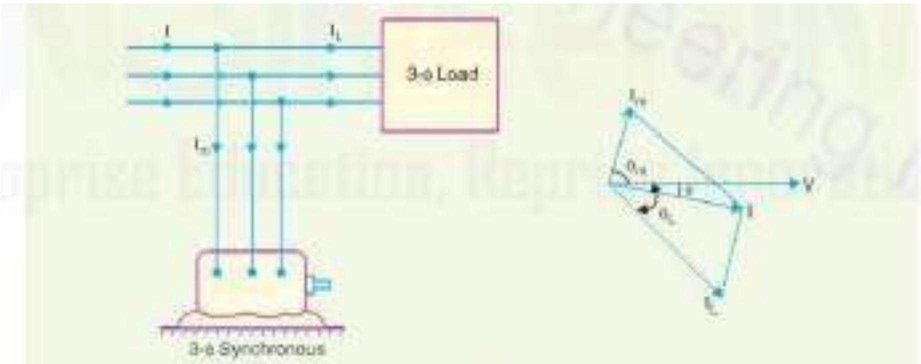


Fig. 6.5

Advantages

- By varying the field excitation, the magnitude of current drawn by the motor can be changed by any amount. This helps in achieving stepless † control of power factor.
- The motor windings have high thermal stability to short circuit currents.
- The faults can be removed easily.

Disadvantages

- There are considerable losses in the motor.
- The maintenance cost is high.
- It produces noise.
- Except in sizes above 500 kVA, the cost is greater than that of static capacitors of the same rating.
- As a synchronous motor hasno self-starting torque, therefore, an auxiliaryequipment has to be provided for this purpose.

Phase advancers. Phase advancers are usedto improve the power factor of induction motors. The low power factor of an induction motor is due to the fact that its stator winding draws exciting current which lags be- hind the supply voltage by 90°. If the exciting ampere turns can be provided from some other a.c. source, then the stator winding willbe relieved of exciting current and the power factor of the motor can be improved. This job is accomplished bythephase advancer which is simply an a.c. exciter. The phase advancer is mounted on the same shaft as the main motor and is connected in the rotor circuit of the motor. It provides exciting ampere turns to the rotor circuit at slip frequency. By providing more ampere turns than required, the induction motor can be made to operate on leading power factor like an over-excited synchronous motor.

Phase advancers have two principal advantages. Firstly, as the exciting ampere turns are supplied at slip frequency, therefore, lagging kVAR drawn by the motor are considerably reduced. Secondly, phase advancer can be conveniently used where the use of synchronous motors is unadmissible. However, the major disadvantage of phase advancers is that they are not economical for motors below 200 H.P.

Load Curves

The curve showing the variation of load on thepower station with respect to (*w.r.t*) *time is known as a* **loadcurve.**

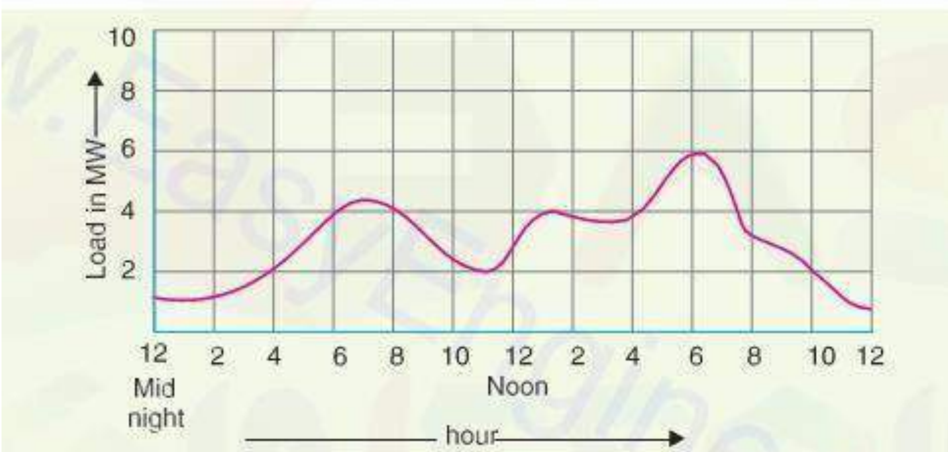


Fig. 3.2

- The daily load curve shows the variations of load on the power station during different hours of the day.
- The area under the daily load curve gives the number of units generated in the day. Units generated/day = Area (in kWh) under daily loadcurve.
- The highest point on the daily load curve represents the maximum demand on the station on that day.
- The area under the daily load curve divided by the total number of hours gives the average load on the station in the day.

$$\text{Average load} = \frac{\text{Area (in kWh) under daily load curve}}{\text{24 hours}}$$

The ratio of the area under the load curve to the total area of rectangle in which it is contained gives the load factor.

$$\text{Load factor} = \frac{\text{Average load}}{\text{Max. demand}} = \frac{\text{Average load} \times 24}{\text{Max. demand} \times 24}$$

$$= \frac{\text{Area (in kWh) under daily load curve}}{\text{Total area of rectangle in which the load curve is contained}}$$

Demand factor.

It is the ratio of maximum demand on the power station to its connected load i.e.

$$\text{Demand factor} = \frac{\text{Maximum demand}}{\text{Connected load}}$$

The value of demand factor is usually less than 1. It is expected because maximum demand on the power station is generally less than the connected load. If the maximum demand on the power station is 80 MW and the connected load is 100 MW, then demand factor = 80/100 = 0·8. The knowledge of demand factor is vital in determining the capacity of the plant equipment.

Maximum demand

It is the greatest demand of load on the power station during a given period.

The load on the power station varies from time to time. The maximum of all the demands that have occurred during a given period (day) is the maximum demand. Thus referring back to the load curve the maximum demand on the power station during the day is 6 MW and it occurs at 6 P.M. Maximum demand is generally less than the connected load because all the consumer do not switch on their connected load to the system at a time. The knowledge of maximum demand is very important as it helps in determining the installed capacity of the station. The station must be capable of meeting the maximum demand.

Load factor

The ratio of average load to the maximum demand during a given period is known as **load factor** *i.e.*,

$$\text{Load factor} = \frac{\text{Average load}}{\text{Max. demand}}$$

If the plant is in operation for T hours,

$$\text{Load factor} = \frac{\text{Average load} \times T}{\text{Max. demand} \times T} = \frac{\text{Units generated in T hours}}{\text{Max. demand} \times \text{T hours}}$$

The load factor may be daily load factor, monthly load factor or annual load factor if the time period considered is a day or month or year. Load factor is always less than 1 because average load is smaller than the maximum demand.

Diversity factor.

The ratio of the sum of individual maximum demands to the maximum demand on power station is known as diversity factor i.e.

$$\text{Diversity factor} = \frac{\text{Sum of individual max. demands}}{\text{Max. demand on power station}}$$

Plant capacity factor

It is the ratio of actual energy produced to the maximum possible energy that could have been produced during a given period i.e.,

$$\text{Plant capacity factor} = \frac{\text{Actual energy produced}}{\text{Max. energy that could have been produced}}$$

$$= \frac{\text{Average demand} \times T \; **}{\text{Plant capacity} \times T}$$

$$= \frac{\text{Average demand}}{\text{Plant capacity}}$$

Connected load

It is the sum of continuous ratings of all the equipments connected to supply system.

A power station supplies load to thousands of consumers. Each consumer has certain equipment installed in his premises. The sum of the continuous ratings of all the equipments in the consumer's premises is the "connected load" of the consumer. For instance, if a consumer has connections of five 100-watt lamps and a power point of 500 watts, then connected load of the consumer is $5 \times 100 + 500 = 1000$ watts. The sum of the connected loads of all the consumers is the connected load to the power station.

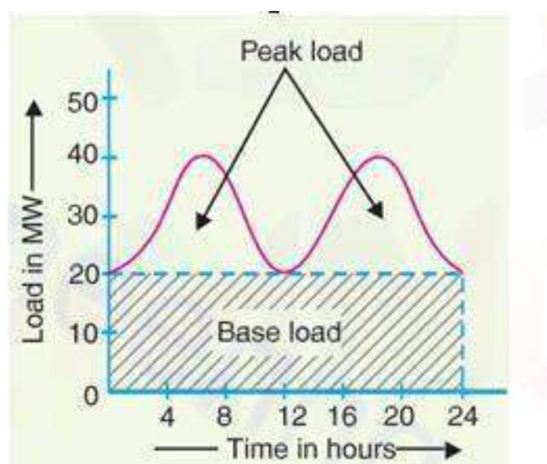


Fig. 3.13

Base Load and Peak Load on Power Station

The changing load on the power station makes its load curve of variable nature. load on the power station varies from time to time. However, load on the power station can be considered in two parts, namely;

(i) Base load
(ii) Peak load

Base load. *The unvarying load which occurs almost the whole day on the station is known as* **base load.**

Peak load. *The various peak demands of load over and above the base load of the station is known as* **peak load.**

Problems:

Example *The maximum demand on a power station is 100 MW. If the annual load factor is 40%, calculate the total energy generated in a year.*

Solution.

$$\text{Energy generated/year} = \text{Max. demand} \times \text{L.F.} \times \text{Hours in a year}$$
$$= (100 \times 10^3) \times (0{\cdot}4) \times (24 \times 365)\ \text{kWh}$$
$$= \mathbf{3504 \times 10^5\ kWh}$$

Example *A generating station has a connected load of 43MW and a maximum demand of 20 MW; the units generated being $61{\cdot}5 \times 10^6$ per annum. Calculate (i) the demand factor and (ii) load factor.*

Solution.

(*i*)
$$\text{Demand factor} = \frac{\text{Max. demand}}{\text{Connected load}} = \frac{20}{43} = \mathbf{0{\cdot}465}$$

(*ii*)
$$\text{Average demand} = \frac{\text{Units generated / annum}}{\text{Hours in a year}} = \frac{61{\cdot}5 \times 10^6}{8760} = 7020\ \text{kW}$$

$$\therefore \quad \text{Load factor} = \frac{\text{Average demand}}{\text{Max. demand}} = \frac{7020}{20 \times 10^3} = \mathbf{0{\cdot}351\ or\ 35{\cdot}1\%}$$

Example *A 100 MW power station delivers 100 MW for 2 hours, 50 MW for 6 hours and is shut down for the rest of each day. It is also shut down for maintenance for 45 days each year. Calculate its annual load factor.*

Solution.

Energy supplied for each working day

$$= (100 \times 2) + (50 \times 6) = 500\ \text{MWh}$$

$$\text{Station operates for} = 365 - 45 = 320 \text{ days in a year}$$

$$\therefore \quad \text{Energy supplied/year} = 500 \times 320 = 160{,}000\ \text{MWh}$$

$$\text{Annual load factor} = \frac{\text{MWh supplied per annum}}{\text{Max. demand in MW} \times \text{Working hours}} \times 100$$
$$= \frac{160{,}000}{(100) \times (320 \times 24)} \times 100 = \mathbf{20{\cdot}8\%}$$

Example *A generating station has a maximum demand of 25MW, a load factor of 60%, a plant capacity factor of 50% and a plant use factor of 72%. Find (i) the reserve capacity of the plant (ii) the daily energy produced and (iii) maximum energy that could be produced daily if the plant while running as per schedule, were fully loaded.*

Solution.

(*i*)
$$\text{Load factor} = \frac{\text{Average demand}}{\text{Maximum demand}}$$

or
$$0{\cdot}60 = \frac{\text{Average demand}}{25}$$

$$\therefore \quad \text{Average demand} = 25 \times 0{\cdot}60 = 15\ \text{MW}$$

$$\text{Plant capacity factor} = \frac{\text{Average demand}}{\text{Plant capacity}}$$

$$\therefore \quad \text{Plant capacity} = \frac{\text{Average demand}}{\text{Plant capacity factor}} = \frac{15}{0{\cdot}5} = 30\ \text{MW}$$

$$\therefore \quad \text{Reserve capacity of plant} = \text{Plant capacity} - \text{maximum demand}$$
$$= 30 - 25 = 5 \text{ MW}$$

(*ii*) $$\text{Daily energy produced} = \text{Average demand} \times 24$$
$$= 15 \times 24 = 360 \text{ MWh}$$

(*iii*) Maximum energy that could be produced

$$= \frac{\text{Actual energy produced in a day}}{\text{Plant use factor}}$$

$$= \frac{360}{0 \cdot 72} = 500 \text{ MWh/day}$$

Example *A diesel station supplies the following loads to various consumers :*

Industrial consumer = 1500 kW ; Commercial establishment = 750 kW

Domestic power = 100 kW; Domestic light = 450 kW

If the maximum demand on the station is 2500 kW and the number of kWh generated per year is 45×10^5*, determine (i) the diversity factor and (ii) annual load factor.*

Solution.

(*i*) $$\text{Diversity factor} = \frac{1500 + 750 + 100 + 450}{2500} = 1 \cdot 12$$

(*ii*) $$\text{Average demand} = \frac{\text{kWh generated / annum}}{\text{Hours in a year}} = 45 \times 10^5/8760 = 513 \cdot 7 \text{ kW}$$

$$\therefore \quad \text{Load factor} = \frac{\text{Average load}}{\text{Max. demand}} = \frac{513 \cdot 7}{2500} = 0 \cdot 205 = 20 \cdot 5\%$$

Example *A power station has a maximum demand of 15000 kW. The annual load factor is 50% and plant capacity factor is 40%. Determine the reserve capacity of the plant.*

Solution.

$$\text{Energy generated/annum} = \text{Max. demand} \times \text{L.F.} \times \text{Hours in a year}$$
$$= (15000) \times (0 \cdot 5) \times (8760) \text{ kWh}$$
$$= 65 \cdot 7 \times 10^6 \text{ kWh}$$

$$\text{Plant capacity factor} = \frac{\text{Units generated / annum}}{\text{Plant capacity} \times \text{Hours in a year}}$$

$$\therefore \quad \text{Plant capacity} = \frac{65 \cdot 7 \times 10^6}{0 \cdot 4 \times 8760} = 18{,}750 \text{ kW}$$

$$\text{Reserve capacity} = \text{Plant capacity} - \text{Max. demand}$$
$$= 18{,}750 - 15000 = 3750 \text{ kW}$$

Chapter9 Types of tariff

Tariff : *The rate at which electrical energy is supplied to a consumer is known as* **tariff**.

Objectives of tariff. Like other commodities, electrical energy is also sold at such a rate so that it not only returns the cost but also earns reasonable profit. Therefore, a tariff should include the following items:

- Recovery of cost of producing electrical energy at the power station.
- Recovery of cost on the capital investment in transmission and distribution systems.
- Recovery of cost of operation and maintenance of supply of electrical energy *e.g.*, metering equipment, billingetc.
- A suitable profit on the capital investment.

Flat rate tariff. *When different types of consumersarecharged at different uniformperunit rates, it is called a* **flat rate tariff**.

In this type of tariff, the consumers are grouped into different classes and each class of consumers is charged at a different uniform rate. For instance, the flat rate per kWh for lighting load may be 60 paise, whereas it may be slightly less† (say 55 paise per kWh) for power load. The different classes of consumers are made taking into account their diversity and load factors. The advantage of such a tariff is that it is more fair to different types of consumers and is quite simple in calculations.

Disadvantages

Since the flat rate tariff varies according to the way the supply is used, separate meters are required for lighting load, power load etc. This makes the application of such a tariff expen- sive and complicated.

A particular class of consumers is charged at the same rate irrespective of the magnitude of energy consumed. However, a big consumer should be charged at a lower rate as in his case the fixed charges per unit arereduced.

Two-part tariff. *When the rate of electrical energy is charged on the basis of maximum demand of the consumer and the units consumed, it is called a* **two-part tariff**.

In two-part tariff, the total charge to be made from the consumer is split into two components *viz*., fixed charges and running charges. The fixed charges depend upon the maximum demand of the consumer while the running charges depend upon the number of units consumed by the consumer. Thus, the consumer is charged at a certain amount per kW of maximum demand plus a certain amount per kWh of energy consumed *i.e.*,

Total charges = Rs ($b \times$ kW + $c \times$ kWh)

where, b = charge per kW of maximumdemand

c = charge per kWh of energy consumed

This type of tariff is mostly applicable to industrial consumers who have appreciable maximum demand.

Advantages

- It is easily understood by the consumers.
- It recovers the fixed charges which depend upon the maximum demand of the consumer but are independent of the units consumed.

Disadvantages

- The consumer has to pay the fixed charges irrespective of the fact whether he has consumed or not consumed the electrical energy.

- There is always error in assessing the maximum demand of the consumer.

Block rate tariff. *When a given block of energy is charged at a specified rate and the succeeding blocks of energy are charged at progressively reduced rates, it is called a* **block rate tariff**.

In block rate tariff, the energy consumption is divided into blocks and the price per unit is fixed in each block. The price per unit in the first block is the highest** and it is progressively reduced for the succeeding blocks of energy. For example, the first 30 units may be charged at the rate of 60 paise per unit ; thenext 25 unitsat therate of 55 paise per unit and theremaining additional units may be charged at the rate of 30 paise per unit.

- The advantage of such a tariff is that the consumer gets an incentive to consume more electrical energy. This increases the load factor of the system and hence the cost of generation is reduced. However, its principal defect is that it lacks a measure of the consumer's demand. This type of tariff is being used for majority of residential and small commercial consumers.

Problems:

Example *A consumer has a maximum demand of 200 kW at 40% load factor. If the tariff is Rs. 100 per kW of maximum demand plus 10 paise per kWh, find the overall cost per kWh.*

Solution.

$$\text{Units consumed/year} = \text{Max. demand} \times \text{L.F.} \times \text{Hours in a year}$$
$$= (200) \times (0{\cdot}4) \times 8760 = 7{,}00{,}800 \text{ kWh}$$

$$\text{Annual charges} = \text{Annual M.D. charges} + \text{Annual energy charges}$$
$$= \text{Rs}(100 \times 200 + 0{\cdot}1 \times 7{,}00{,}800)$$
$$= \text{Rs } 90{,}080$$

$$\therefore \quad \text{Overall cost/kWh} = \text{Rs } \frac{90{,}080}{7{,}00{,}800} = \text{Re } 0{\cdot}1285 = 12{\cdot}85 \text{ paise}$$

Example *The maximum demand of a consumer is 20 A at 220 V and his total energy consumption is 8760 kWh. If the energy is charged at the rate of 20 paise per unit for 500 hours use of the maximum demand per annum plus 10 paise per unit for additional units, calculate : (i) annual bill (ii) equivalent flat rate.*

Solution.

Assume the load factor and power factor to be unity.

$$\therefore \quad \text{Maximum demand} = \frac{220 \times 20 \times 1}{1000} = 4{\cdot}4 \text{ kW}$$

(*i*) $\text{Units consumed in 500 hrs} = 4{\cdot}4 \times 500 = 2200 \text{ kWh}$

$$\text{Charges for 2200 kWh} = \text{Rs } 0{\cdot}2 \times 2200 = \text{Rs } 440$$
$$\text{Remaining units} = 8760 - 2200 = 6560 \text{ kWh}$$

$$\text{Charges for 6560 kWh} = \text{Rs } 0{\cdot}1 \times 6560 = \text{Rs } 656$$

$$\therefore \quad \text{Total annual bill} = \text{Rs } (440 + 656) = \textbf{Rs. 1096}$$

(*ii*) $$\text{Equivalent flat rate} = \text{Rs } \frac{1096}{8760} = \text{Re } 0{\cdot}125 = \textbf{12.5 paise}$$

Example *The following two tariffs are offered :*

(*a*) *Rs 100 plus 15 paise per unit ;*

(*b*) *A flat rate of 30 paise per unit ;*

At what consumption is first tariff economical ?

Solution.

Let x be the number of units at which charges due to both tariffs become equal. Then,

$$100 + 0{\cdot}15x = 0{\cdot}3x$$

or $$0{\cdot}15x = 100$$

$\therefore$ $$x = 100/0{\cdot}15 = \textbf{666·67 units}$$

Therefore, tariff (*a*) is economical if consumption is more than 666·67 units.

Example . *A supply is offered on the basis of fixed charges of Rs 30 per annum plus 3 paise per unit or alternatively, at the rate of 6 paise per unit for the first 400 units per annum and 5 paise per unit for all the additional units. Find the number of units taken per annum for which the cost under the two tariffs becomes the same.*

Solution. Let x (> 400) be the number of units taken per annum for which the annual charges due to both tariffs become equal.

$$\text{Annual charges due to first tariff} = \text{Rs } (30 + 0{\cdot}03x)$$

$$\text{Annual charges due to second tariff} = \text{Rs } [(0{\cdot}06 \times 400) + (x - 400) \times 0{\cdot}05]$$

$$= \text{Rs } (4 + 0{\cdot}05x)$$

As the charges in both cases are equal,

$\therefore$ $$30 + 0{\cdot}03x = 4 + 0{\cdot}05x$$

or $$x = \frac{30 - 4}{0{\cdot}05 - 0{\cdot}03} = \textbf{1300 kWh}$$

Example *An electric supply company having a maximum load of 50 MW generates 18×10^7 units per annum and the supply consumers have an aggregate demand of 75 MW. The annual expenses including capital charges are :*

For fuel = Rs 90 lakhs

Fixed charges concerning generation = Rs 28 lakhs

Fixed charges concerning transmission and distribution = Rs 32 lakhs

Assuming 90% of the fuel cost is essential to running charges and the loss in transmission and distribution as 15% of kWh generated, deduce a two part tariff to find the actual cost of supply to the consumers.

Solution.

Annual fixed charges

$$\text{For generation} = \text{Rs } 28 \times 10^5$$

$$\text{For transmission and distribution} = \text{Rs } 32 \times 10^5$$

$$\text{For fuel (10\% only)} = \text{Rs } 0{\cdot}1 \times 90 \times 10^5 = \text{Rs } 9 \times 10^5$$

$$\text{Total annual fixed charge} = \text{Rs } (28 + 32 + 9) \times 10^5 = \text{Rs } 69 \times 10^5$$

This cost has to be spread over the aggregate maximum demand of all the consumers *i.e.*, 75 MW.

$$\therefore \quad \text{Cost per kW of maximum demand} = \text{Rs}\,\frac{69 \times 10^5}{75 \times 10^3} = \textbf{Rs. 92}$$

Annual running charges.

$$\text{Cost of fuel (90\%)} = \text{Rs}\,0{\cdot}9 \times 90 \times 10^5 = \text{Rs}\,81 \times 10^5$$

$$\text{Units delivered to consumers} = 85\%\text{ of units generated}$$

$$= 0{\cdot}85 \times 18 \times 10^7 = 15{\cdot}3 \times 10^7\,\text{kWh}$$

This cost is to be spread over the units delivered to the consumers.

$$\therefore \quad \text{Cost/kWh} = \text{Rs}\,\frac{81 \times 10^5}{15{\cdot}3 \times 10^7} = \text{Re}\,0{\cdot}053 = \textbf{5.3 paise}$$

$\therefore$ Tariff is Rs 92 per kW of maximum demand plus 5·3 paise per kWh.

Example *A generating station has a maximum demand of 75 MW and a yearly load factor of 40%. Generating costs inclusive of station capital costs are Rs. 60 per annum per kW demand plus 4 paise per kWh transmitted. The annual capital charges for transmission system are Rs 20,00,000 and for distribution system Rs 15,00,000 ; the respective diversity factors being 1·2 and 1·25. The efficiency of transmission system is 90% and that of the distribution system inclusive of substation losses is 85%. Find the yearly cost per kW demand and cost per kWh supplied :*

(*i*) *at the substation* (*ii*) *at the consumers premises.*

Solution.

$$\text{Maximum demand} = 75\,\text{MW} = 75{,}000\,\text{kW}$$

$$\text{Annual load factor} = 40\% = 0{\cdot}4$$

(*i*) **Cost at substation.** The cost per kW of maximum demand is to be determined from the total annual fixed charges associated with the supply of energy at the substation. The cost per kWh shall be determined from the running charges.

(*a*) **Annual fixed charges**

$$\text{Generation cost} = \text{Rs}\,60 \times 75 \times 10^3 = \text{Rs}\,4{\cdot}5 \times 10^6$$

$$\text{Transmission cost} = \text{Rs}\,2 \times 10^6$$

Total annual fixed charges at the substation

$$= \text{Rs}\,(4{\cdot}5 + 2) \times 10^6 = \text{Rs}\,6{\cdot}5 \times 10^6$$

Aggregate of all maximum demands by the various substations

$$= \text{Max. demand on generating station} \times \text{Diversity factor}$$

$$= (75 \times 10^3) \times 1{\cdot}2 = 90 \times 10^3\,\text{kW}$$

The total annual fixed charges have to be spread over the aggregate maximum demands by various substations *i.e.*, 90×10^3 kW.

Annual cost per kW of maximum demand

$$= \text{Rs}\,\frac{6{\cdot}5 \times 10^6}{90 \times 10^3} = \textbf{Rs. 72.22}$$

(*b*) **Running Charges.** It is given that cost of 1 kWh transmitted to substation is 4 paise. As the transmission efficiency is 90%, therefore, for every kWh transmitted, 0·9 kWh reaches the substation.

$$\therefore \quad \text{Cost/kWh at substation} = 4/0{\cdot}9 = \textbf{4·45 paise}$$

Hence at sub-station, the cost is **Rs 72·22** per annum per kW maximum demand plus **4·45 paise** per kWh.

(*ii*) Cost at consumer's premises. The total annual fixed charges at consumer's premises is the sum of annual fixed charges at substation (*i.e.* Rs $6{\cdot}5\times 10^6$) and annual fixed charge for distribution (*i.e.*, Rs $1{\cdot}5\times 10^6$).

∴ Total annual fixed charges at consumer's premises

$$= \text{Rs}\,(6{\cdot}5 + 1{\cdot}5)\times 10^6 = \text{Rs}\,8\times 10^6$$

Aggregate of maximum demands of all consumers

$$= \text{Max. demand on Substation} \times \text{Diversity factor}$$

$$= (90\times 10^3)\times 1{\cdot}25 = 112{\cdot}5\times 10^3\ \text{kW}$$

∴ Annual cost per kW of maximum demand

$$= \text{Rs}\,\frac{8\times 10^6}{112\cdot 5\times 10^3} = \textbf{Rs. 71.11}$$

As the distribution efficiency is 85%, therefore, for each kWh delivered from substation, only 0·85 kWh reaches the consumer's premises.

∴ Cost per kWh at consumer's premises

$$= \frac{\text{Cost per kWh at substation}}{0\cdot 85} = \frac{4\cdot 45}{0\cdot 85} = \textbf{5.23 paise}$$

Hence at consumer's premises, the cost is **Rs. 71·11** per annum per kW maximum demand plus **5·23 paise** per kWh.

Chapter10 substation

Substation:

The assembly of apparatus used to change some characteristic (e.g. voltage, a.c. to d.c., frequency,

p.f. etc.) of electric supply is called a **sub-station.**

Sub-stations are important part of power system. The continuity of supply depends to a considerable extent upon the successful operation of substations. It is, therefore, essential to exercise utmost care while designing and building a sub-station. The following are the important points which must be kept in view while laying out a substation :

- It should be located at a proper site. As far as possible, it should be located at the centre of gravity of load.
- It should provide safe and reliable arrangement. For safety, consideration must be given to the maintenance of regulation clearances, facilities for carrying out repairs and maintenance, abnormal occurrences such as possibility of explosion or fire etc. For reliability, consideration must be given for good design and construction, the provision of suitable protective gear *etc.*
- It should be easily operated and maintained.
- It should involve minimum capital cost

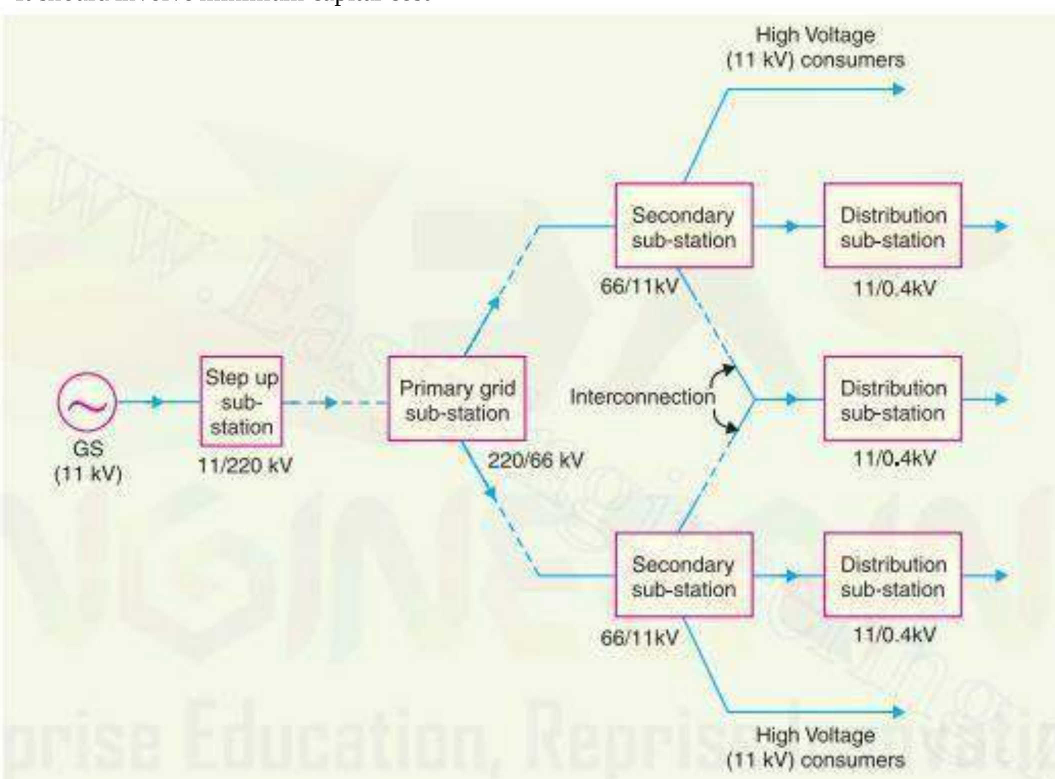
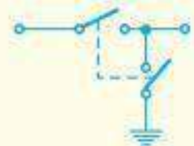
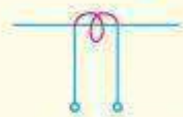
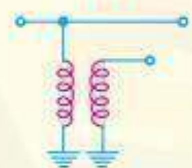
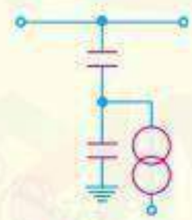
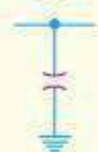
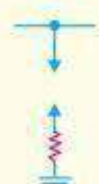
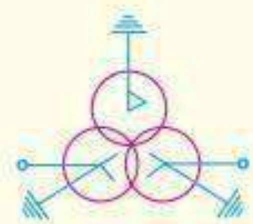
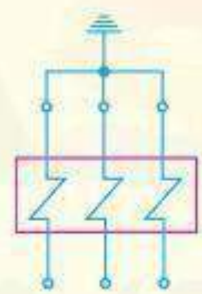
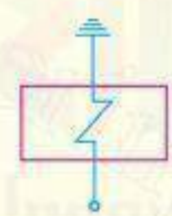


Fig. 25.1

Symbols for equipment in substation:

S.No.	Circuit element	Symbol
1	Bus-bar	
2	Single-break isolating switch	
3	Double-break isolating switch	
4	On load isolating switch	
5	Isolating switch with earth Blade	
6	Current transformer	
7	Potential transformer	
8	Capacitive voltage transformer	
9	Oil circuit breaker	
10	Air circuit breaker with overcurrent tripping device	
11	Air blast circuit breaker	
12	Lightning arrester (active gap)	
13	Lightning arrester (valve type)	

S.No.	Circuit element	Symbol
14	Arcing horn	
15	3-ϕ Power transformer	
16	Overcurrent relay	
17	Earth fault relay	

Key diagram of 66/11 KV substation:

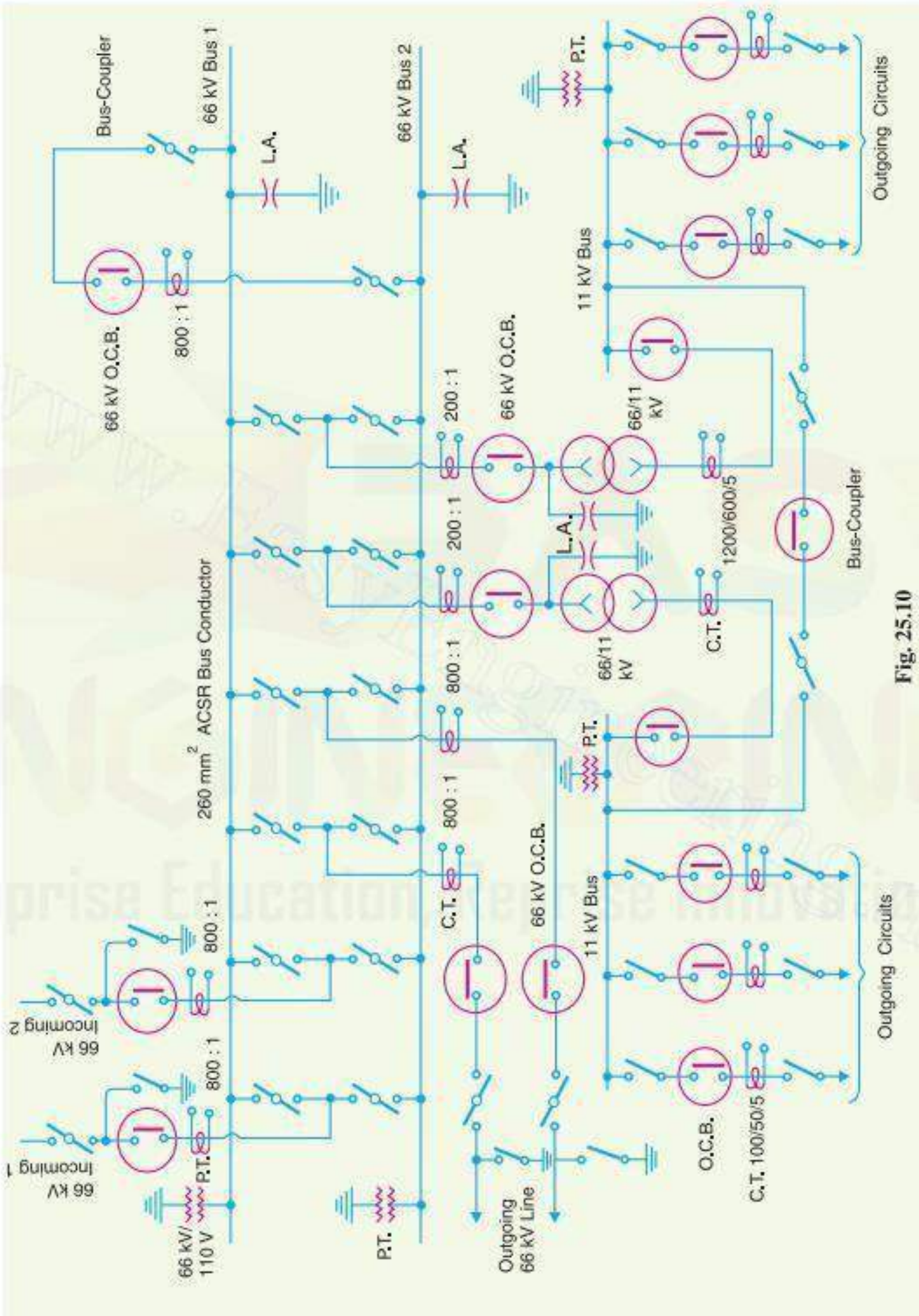


Fig. 25.10

- There are two 66 kV incoming lines marked ‘incoming 1’ and ‘incoming 2’ connected to the bus-bars. Such an arrangement of two incoming lines is called a double circuit. Each incoming line is capable of supplying the rated sub-station load. Both these lines can be loaded simultaneously to share the sub-station load or any one line can be called upon to meet the entire load. The double circuit arrangement increases the reliability of the system. In case there is a breakdown of one incoming line, the continuity of supply can be maintained by the other line.
- The sub-station has duplicate bus-bar system; one ‘main bus-bar’ and the other spare bus- bar. The incoming lines can be connected to either bus-bar with the help of a bus-coupler which consists of a circuit breaker and isolators. The advantage of double bus-bar system is that if repair is to be carried on one bus-bar, the supply need not be interrupted as the entire load can be transferred to the other bus.
- There is an arrangement in the sub-station by which the same 66 kV double circuit supply is going out *i.e.* 66 kV double circuit supply is passing through the sub-station. The outgoing 66 kV double circuit line can be made to act as incoming line.
- There is also an arrangement to step down the incoming 66 kV supply to 11 kV by two units of 3-phase transformers; each transformer supplying to a separate bus-bar. Generally, one transformer suppliesthe entire sub-station load whilethe other transformer acts as a standby unit. If need arises, both the transformers can be called upon to share the sub-station load. The 11 kV outgoing lines feed to the distribution sub-stations located near consumers localities.
- Both incoming and outgoing lines are connected through circuit breakers having isolators on their either end. Whenever repair is to be carried over the line towers, the line is first switched off and then earthed.
- The potential transformers (P.T.) and current transformers (C.T.) and suitably located for supply to metering and indicating instruments and relay circuits (not shown in the figure). The P.T. is connected right on the point where the line is terminated. The CTs are connected at the terminals of each circuit breaker.
- The lightning arresters are connected near the transformer terminals (on H.T. side) to protect them from lightningstrokes.
- There are other auxiliary components in the sub-station such as capacitor bank for power factor improvement, earth connections, local supply connections, d.c. supply connections etc. However, these have been omitted in the key diagram for the sake of simplicity.

Key diagram of 11KV/400V substation:

- The 3-phase, 3-wire 11 kV line is tapped and brought to the gang operating switch installed near the sub-station. The G.O. switch consists of isolators connected in each phase of the 3- phase line.
- From the G.O. switch, the 11 kV line is brought to the indoor sub-station as underground cable. It is fed to the H.T. side of the transformer (11 kV/400 V) *via* the 11 kV O.C.B. The transformer steps down the voltage to 400 V, 3-phase, 4-wire.
- The secondary of transformer supplies to the bus-bars *via* the main O.C.B. From the bus- bars, 400 V, 3-phase, 4-wire supply is given to the various consumers *via* 400 V O.C.B. The voltage between any two phases is 400 V and between any phase and neutral it is 230 V. The single phase residential load is connected between any one phase and neutral whereas 3- phase, 400 V motor load is connected across 3-phase lines directly.
- The CTs are located at suitable places in the sub-station circuit and supply for the metering and indicating instruments and relaycircuits.

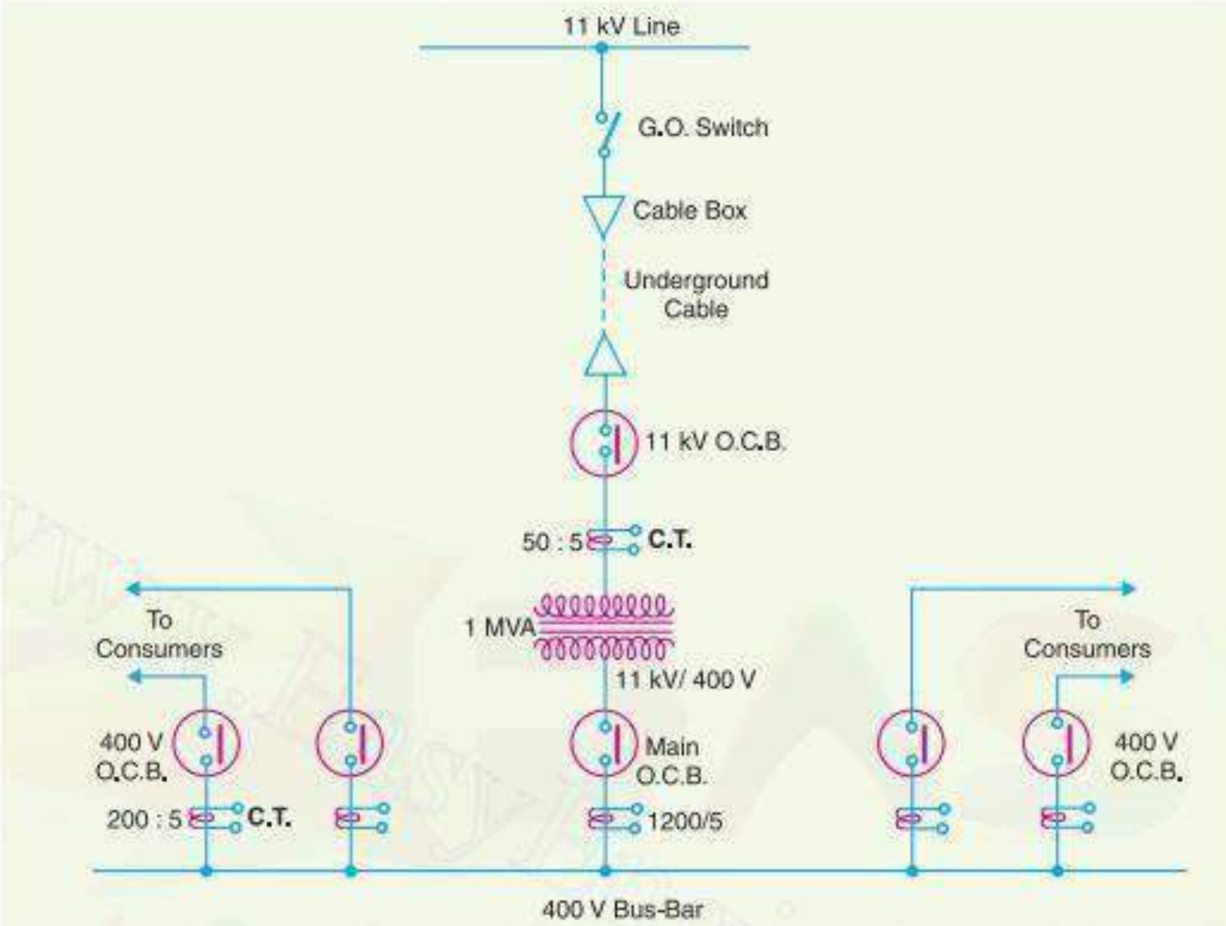


Fig. 25.11

Earthing or Grounding:

The process of connecting the metallic frame (i.e. non-current carrying part) of electrical equipment or some electrical part of the system (e.g. neutral point in a star-connected system, one conductor of the secondary of a transformer etc.) to earth (i.e. soil) is called **grounding** *or* **earthing.**

If grounding is done systematically in the line of the power system, we can effectively prevent accidents and damage to the equipment of the power system and at the same time continuity of supply can be maintained. Grounding or earthing maybe classified as : ***(i)*** Equipment grounding (ii) System grounding.

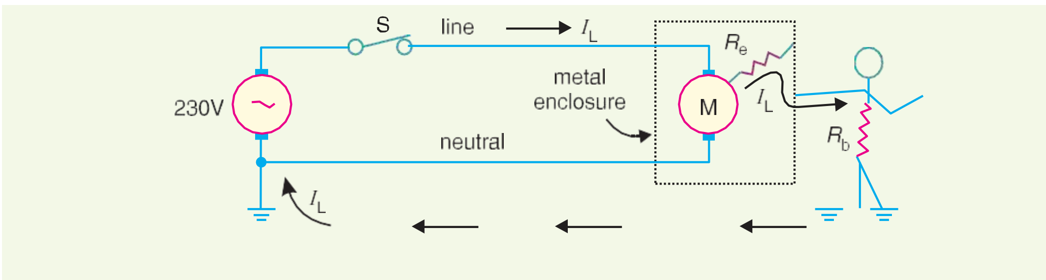
Equipment grounding deals with earthing the non-current-carrying metal parts of the electrical equipment. On the other hand, system grounding means earthing some part of the electrical system *e.g.* earthing of neutral point of star-connected system in generating stations and sub-stations.

Equipment grounding:

The process of connecting non-current-carrying metal parts (i.e. metallic enclosure) of the electrical equipment to earth (i.e. soil) in such a way that in case of insulation failure, the enclosure effectively remains at earth potential is called **equipment grounding.**

We are frequently in touch with electrical equipment of all kinds, ranging from domestic appliances and hand-held tools to industrial motors. We shall illustrate the need of effective equipment grounding by considering a single-phase circuit composed of a 230 V source connected to a motor M as shown in Fig. 26.1. Note that neutral is solidly grounded at the service entrance. In the interest of easy understanding, we shall divide the discussion into three heads *viz.* **(i)** Ungrounded enclosure (ii) enclosure connected to neutral wire ***(iii)*** ground wire connected to enclosure.

- **Ungrounded enclosure.** Fig. 26.1 shows the case of ungrounded metal enclosure. If a person touches the metal enclosure, nothing will happen if the equipment is functioning correctly. But if the winding insulation becomes faulty, the resistance R_e between the motor and enclosure drops to a low value (a few hundred ohms or less). A person having a body resistance R_b would complete the current path as shown in Fig. 26.1.



If R_e is small (as is usually the case when insulation failure of winding occurs), the leakage current I_L through the person's body could be dangerously high. As a result, the person would get severe *electric shock which may be fatal. Therefore, this system is unsafe.

- **Enclosure connected to neutral wire.** It may appear that the above problem can be solved by connecting the enclosure to the grounded neutral. Now the leakage current I_L flows from the motor, through the enclosure and straight back to the neutral wire . Therefore, the enclosure remains at earth potential. Consequently, the operator would not experience any electric shock.

The trouble with this method is that the neutral wire may become open either accidentally or due to a faulty installation. For example, if the switch is inadvertently in series with the neutral rather than the live wire the motor can still be turned on and off. However, if someone touched the enclosure while the motor is *off,* he would receive a severe electric shock. It is because when the motor is off, the potential of the enclosure rises to that of the live conductor.

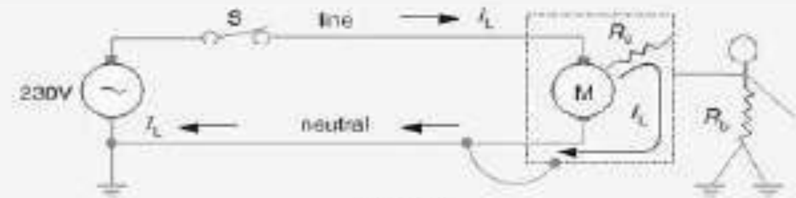


Fig. 26.2

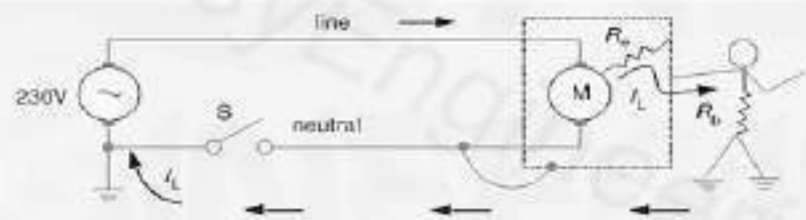


Fig. 26.3

- **Ground wire connected to enclosure.** To get rid of this problem, we install a third wire, called *ground wire*, between the enclosure and the system ground as shown in Fig. 26.4. The ground wire may be bare or insulated. If it is insulated, it is coloured green. Electrical outlets have three contacts — one for live wire, one for neutral wire and one for ground wire.

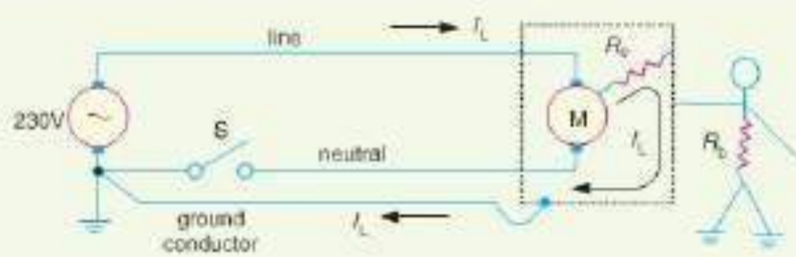


Fig. 26.4

System Grounding:

The process of connecting some electrical part of the power system (e.g. neutral point of a star-connected system, one conductor of the secondary of a transformer etc.) to earth (i.e. soil) is called **system grounding.**

- the primary winding of a distribution transformer connected between the line and neutral of a 11 kV line. If the secondary conductors are *ungrounded*, it would appear that a person could touch either secondary conductor without harm because there is no ground return. However, this is not true. Referring to Fig. 26.5, there is capacitance C_1 between primary and secondary and capacitance C_2 between secondary and ground. This capacitance coupling can produce a high voltage between the secondary lines and the ground. Depending upon the relative magnitudes of C_1 and C_2, it may be as high as 20% to 40% of the primary voltage. If a person touches either one of the secondary wires, the resulting capacitive current I_C flowing through the body could be dangerous even in case of small transformers. For example, if I_C is only 20 mA, the person may get a fatal electric shock. If one of the secondary conductors is grounded, the capacitive coupling almost reduces to zero and so is the capacitive current I_C. As a result, the person will experience no electric shock. This explains the importance of system grounding.

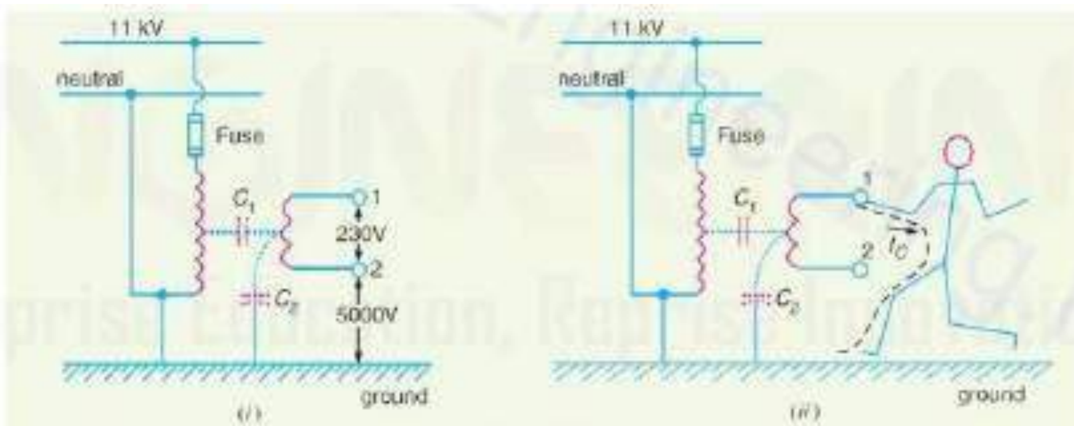


Fig. 26.5

- Let the primary winding of a distribution transformer connected between the line and neutral of a 11 kV line. The secondary conductors are ungrounded. Suppose that the high voltage line (11 kV in thiscase) touches the 230 V conductor. This could be caused by an internal fault in the transformer or by a branch or tree falling across the 11 kV and 230 V lines. Under these circumstances, a very high voltage is imposed between the secondary conductors and ground. This would immediately puncture the 230 V insulation, causing a massive flashover. This flashover could occur anywhere on the secondary network, possiblyinside a home or factory. Therefore, ungrounded secondary in this case is a potential fire hazard and may produce grave accidents under abnormal conditions If one of the secondary lines is grounded, the accidental contact between a 11 kV conductor and a 230 V conductor produces a dead short. The short-circuit current (*i.e.* fault current) follows the dotted path. This large current will blow the fuse on the 11 kV side, thus disconnecting the transformer and secondary distribution system from the 11 kV line. This explains the importance of system grounding in the line of the power system.

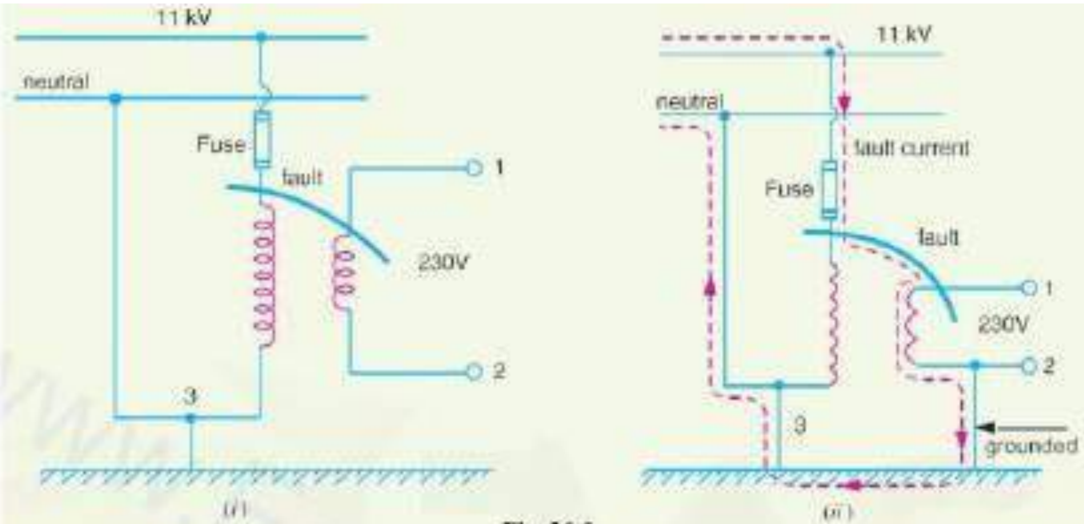

11 kV
neutral
Fuse
fault
1
230V
2
3
(i)
11 kV
neutral
fault current
Fuse
fault
1
230V
2
3
grounded
(ii)

Fig. 26.6