

ANNAMACHARYA **INSTITUTE OF TECHNOLOGY AND SCIENCES** **(AUTONOMOUS)**

Approved by AICTE, New Delhi & Permanent Affiliation to JNTUA, Anantapur.

Three B. Tech Programmes (CSE , ECE & CE) are accredited by NBA, New Delhi,Accredited by NAAC with 'A' Grade , Bangalore.

A-grade awarded by AP Knowledge Mission. Recognized under sections 2(f) & 12(B) of UGC Act 1956.

Venkatapuram Village, Renigunta Mandal, Tirupati, Andhra Pradesh-517520.

Department of Computer Science and Engineering



Academic Year 2023-24

II. B.Tech I Semester

**Basics of Electrical and Electronics
Engineering**

(Common to CSE,CIC,CSE(DS))

(20AES0205)

Prepared By

Mr BALAJI NANDA KUMAR

Mr P ANIL KUMAR

Assistant Professor

Department of EEE ,ECE, AITS

UNIT – I DC & AC CIRCUITS

Basic Fundamental Concepts

Concept of Charge

In all the atoms, there exists number of electrons which are very loosely bound to its nucleus. Such electrons are free to wander about, through the space under the influence of specific forces. Now when such electrons are removed from an atom it becomes positively charged. This is because of losing negatively charged particles i.e. electrons from it. As against this, if excess electrons are added to the atom it becomes negatively charged.

Key Point: Thus total deficiency or addition of excess electrons in an atom is called its charge and the element is said to be charged.

The following table shows the different particles and charge possessed by them.

Particle	Charge possessed in Coulomb	Nature
Neutron	0	Neutral
Proton	1.602×10^{-19}	Positive
Electron	1.602×10^{-19}	Negative

Unit of Charge

As seen from the Table 1.2 that the charge possessed by the electron is very very small hence it is not convenient to take it as the unit of charge.

The unit of the measurement of the charge is Coulomb.

The charge on one electron is 1.602×10^{-19} , so one coulomb charge is defined as the charge possessed by total number of $(1 / 1.602 \times 10^{-19})$ electrons i.e. 6.24×10^{18} number of electrons.

Thus, $1 \text{ coulomb} = \text{charge on } 6.24 \times 10^{18} \text{ electrons}$

From the above discussion it is clear that if an element has a positive charge of one coulomb then that element has a deficiency of 6.24×10^{18} number of electrons.

Key Point: Thus, addition or removal of electrons causes the change in the nature of the charge possessed by the element.

Relation between Charge and Current

The current is flow of electrons. Thus current can be measured by measuring how many electrons are passing through material per second. This can be expressed in terms of the charge carried by those electrons in the material per second. So the flow of charge per unit time is used to quantify an electric current.

Key Point: So current can be defined as rate of flow of charge in an electric circuit or in any medium in which charges are subjected to an external electric field.

The charge is indicated by **Q** coulombs while current is indicated by **I**. The unit for the current is **Amperes** which is nothing but coulombs/sec. Hence mathematically we can write the relation between the charge (**Q**) and the electric current (**I**) as,

$$I = \frac{Q}{t} \text{ Amperes}$$

where
 I = Average current flowing
 Q = Total charge transferred
 t = Time required for transfer of charge.

Definition of 1 Ampere : A current of 1 Ampere is said to be flowing in the conductor when a charge of one coulomb is passing any given point on it in one second.

Now 1 coulomb is 6.24×10^{18} number of electrons. So 1 ampere current flow means flow of 6.24×10^{18} electrons per second across a section taken any where in the circuit.

$$1 \text{ Ampere current} = \text{Flow of } 6.24 \times 10^{18} \text{ electrons per second}$$

Concept of Electric Potential and Potential Difference

When two similarly charged particles are brought near, they try to repel each other while dissimilar charges attract each other. This means, every charged particle has a tendency to do work.

Key Point: This ability of a charged particle to do the work is called its *electric potential*. The unit of electric potential is *volt*.

The electric potential at a point due to a charge is one volt if one joule of work is done in bringing a unit positive charge i.e. positive charge of one coulomb from infinity to that point.

Mathematically it is expressed as,

$$\text{Electrical Potential} = \frac{\text{Workdone}}{\text{Charge}} = \frac{W}{Q}$$

Let us define now the potential difference.

It is well known that, flow of water is always from higher level to lower level, flow of heat is always from a body at higher temperature to a body at lower temperature. Such a level difference which causes flow of water, heat and so on, also exists in electric circuits. In electric circuits flow of current is always from higher electric potential to lower electric potential. So we can define potential difference as below :

Key Point : The difference between the electric potentials at any two given points in a circuit is known as *Potential Difference (p.d.)*. This is also called *voltage* between the two points and measured in volts. The symbol for voltage is V .

For example, let the electric potential of a charged particle A is say V_1 while the electric potential of a charged particle B is say V_2 . Then the potential difference between the two particles A and B is $V_1 - V_2$. If $V_1 - V_2$ is positive we say that A is at higher potential than B while if $V_1 - V_2$ is negative we say that B is at higher potential than A.

Key Point: The potential difference between the two points is one volt if one joule of work is done in displacing unit charge (1 coulomb) from a point of lower potential to a point of higher potential.

Consider two points having potential difference of V volts between them, as shown in the Fig. 1.4. The point A is at higher potential than B. As per the definition of volt, the V joules of work is to be performed to move unit charge from point B to point A.

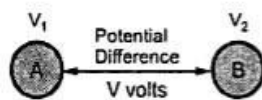


Fig. 1.4

Thus, when such two points, which are at different potentials are joined together with the help of wire, the electric current flows from higher potential to lower potential i.e. the electrons start flowing from lower potential to higher potential. Hence, to maintain the flow of electrons i.e. flow of electric current, there must exist a potential

difference between the two points.

Key Point: No current can flow if the potential difference between the two points is zero.

Electric Current

The directed flow of free electrons (or charge) is called **electric current**. The flow of electric current can be beautifully explained by referring to Fig. 1.1. The copper strip has a large number of free electrons. When electric pressure or voltage is applied, then free electrons, being negatively charged, will start moving towards the positive terminal around the circuit as shown in Fig. 1.1. This directed flow of electrons is called electric current.

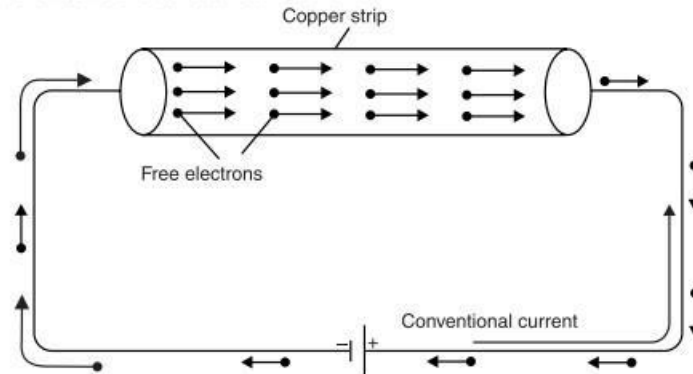


Fig. 1.1

The reader may note the following points :

- (i) Current is flow of electrons and electrons are the constituents of matter. Therefore, electric current is matter (*i.e.* free electrons) in motion.
- (ii) The actual direction of current (*i.e.* flow of electrons) is from negative terminal to the positive terminal through that part of the circuit external to the cell. However, prior to Electron theory, it was assumed that current flowed from positive terminal to the negative terminal of the cell

Types of Electric Current

The electric current may be classified into three main classes: (i) steady current (ii) varying current and (iii) alternating current.

(i) **Steady current.** When the magnitude of current does not change with time, it is called a steady current. Fig. 1.3 (i) shows the graph between steady current and time. Note that value of current remains the same as the time changes. The current provided by a battery is almost a steady current (*d.c.*).

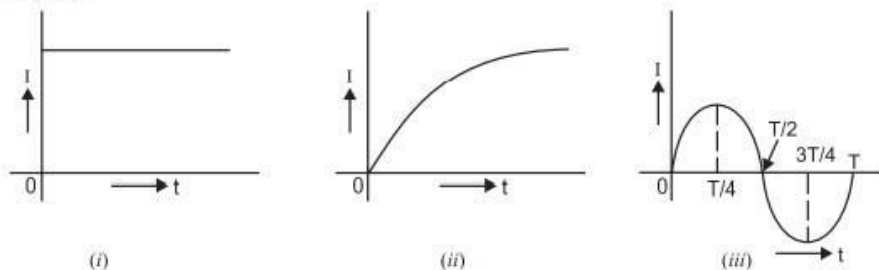


Fig. 1.3

(ii) **Varying current.** When the magnitude of current changes with time, it is called a varying current. Fig. 1.3 (ii) shows the graph between varying current and time. Note that value of current varies with time.

(iii) **Alternating current.** An alternating current is one whose magnitude changes continuously with time and direction changes periodically. Due to technical and economical reasons, we produce alternating currents that have sine waveform (or cosine waveform) as shown in Fig. 1.3 (iii). It is called *alternating current* because current flows in alternate directions in the circuit, *i.e.*, from 0 to $T/2$ second (T is the time period of the wave) in one direction and from $T/2$ to T second in the opposite direction. The current provided by an a.c. generator is alternating current that has sine (or cosine) waveform.

via the circuit. This convention is so firmly established that it is still in use. This assumed direction of current is now called *conventional current*.

Unit of Current. The strength of electric current I is the rate of flow of electrons *i.e.* charge flowing per second.

$$\therefore \text{Current, } I = \frac{Q}{t}$$

The charge Q is measured in coulombs and time t in seconds. Therefore, the unit of electric current will be *coulombs/sec or ampere*. If $Q = 1$ coulomb, $t = 1$ sec, then $I = 1/1 = 1$ ampere.

One ampere of current is said to flow through a wire if at any cross-section one coulomb of charge flows in one second.

Thus, if 5 amperes current is flowing through a wire, it means that 5 coulombs per second flow past any cross-section of the wire.

Note. 1 C = charge on 625×10^{16} electrons. Thus when we say that current through a wire is 1 A, it means that 625×10^{16} electrons per second flow past any cross-section of the wire.

$$\therefore I = \frac{Q}{t} = \frac{ne}{t} \quad \text{where } e = -1.6 \times 10^{-19} \text{ C ; } n = \text{number of electrons}$$

Concept of E.M.F. and Potential Difference

There is a distinct difference between *e.m.f.* and potential difference. The *e.m.f.* of a device, say a battery, is a measure of the energy the battery gives to each coulomb of charge. Thus if a battery supplies 4 joules of energy per coulomb, we say that it has an *e.m.f.* of 4 volts. The energy given to each coulomb in a battery is due to the chemical action.

The potential difference between two points, say A and B , is a measure of the energy used by one coulomb in moving from A to B . Thus if potential difference between points A and B is 2 volts, it means that each coulomb will give up an energy of 2 joules in moving from A to B .

Illustration. The difference between *e.m.f.* and *p.d.* can be made more illustrative by referring to Fig. 1.8. Here battery has an *e.m.f.* of 4 volts. It means battery supplies 4 joules of energy to each coulomb continuously. As each coulomb travels from the positive terminal of the battery, it gives up its most of energy to resistances (2Ω and 2Ω in this case) and remaining to connecting wires. When it returns to the negative terminal, it has lost all its energy originally supplied by the battery. The battery now supplies fresh energy to each coulomb (4 joules in the present case) to start the journey once again.

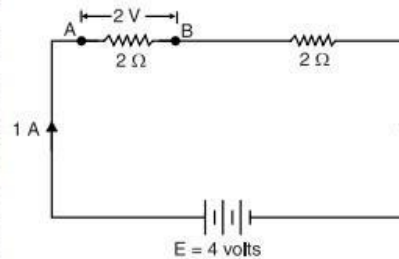


Fig. 1.8

The *p.d.* between any two points in the circuit is the energy used by one coulomb in moving from one point to another. Thus in Fig. 1.8, *p.d.* between A and B is 2 volts. It means that 1 coulomb will give up an energy of 2 joules in moving from A to B . This energy will be released as heat from the part AB of the circuit.

The following points may be noted carefully :

- (i) The name *e.m.f.* at first sight implies that it is a force that causes current to flow. This is not correct because it is not a force but energy supplied to charge by some active device such as a battery.
- (ii) *Electromotive force (e.m.f.) maintains potential difference while p.d. causes current to flow.*

Resistance

The opposition offered by a substance to the flow of electric current is called its **resistance**.

Since current is the flow of free electrons, resistance is the opposition offered by the substance to the flow of free electrons. This opposition occurs because atoms and molecules of the substance obstruct the flow of these electrons. Certain substances (*e.g.* metals such as silver, copper, aluminium etc.) offer very little opposition to the flow of electric current and are called conductors. On the other hand, those substances which offer high opposition to the flow of electric current (*i.e.* flow of free electrons) are called insulators *e.g.* glass, rubber, mica, dry wood etc.

It may be noted here that resistance is the electric friction offered by the substance and causes production of heat with the flow of electric current. The moving electrons collide with atoms or molecules of the substance ; each collision resulting in the liberation of minute quantity of heat.

Unit of resistance. The practical unit of resistance is ohm and is represented by the symbol Ω . It is defined as under :

A wire is said to have a resistance of **1 ohm** if a *p.d.* of 1 volt across its ends causes 1 ampere to flow through it (See Fig. 1.10).

There is another way of defining ohm.

A wire is said to have a resistance of **1 ohm** if it releases 1 joule (or develops 0.24 calorie of heat) when a current of 1 A flows through it for 1 second.

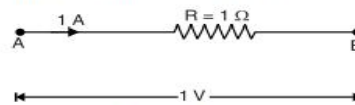


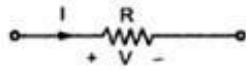
Fig. 1.10

Electrical Circuit Elements[R-L-C] Parameters

Basic Circuit Components

Let us take a brief review of three basic elements namely resistance, capacitance and inductance.

1 Resistance



It is the property of the material by which it opposes the flow of current through it. The resistance of element is denoted by the symbol 'R'. Resistance is measured in ohms (Ω).

The resistance of a given material depends on the physical properties of that material and given by,

$$R = \frac{\rho l}{a}$$

where

- l = Length in metres
- a = Cross-sectional area in square metres
- ρ = Resistivity in ohms-metres
- R = Resistance in ohms

We can define unit ohm as below.

Key Point: 1 Ohm : The resistance of a circuit, in which a current of 1 ampere generates the heat at the rate of 1 joules per second is said to be 1 ohm.

Now $4.186 \text{ Joules} = 1 \text{ Calorie}$
 hence $1 \text{ Joule} = 0.24 \text{ Calorie}$

Thus unit 1 ohm can be defined as that resistance of the circuit if it develops 0.24 calories of heat, when one ampere current flows through the circuit for one second.

The unit ohm also can be defined as, one ohm resistance is that which allows one ampere current to flow through it when one volt voltage is impressed across it.

The relation between voltage and current for a resistance is given by **Ohm's law** as,

$$v = R i$$

$$\therefore R = \frac{v}{i}$$

The power absorbed by a resistance is given by,

$$\therefore p(t) = vi = \frac{v^2}{R} = i^2 R \text{ watts}$$

while the amount of energy converted to heat energy in time t is given by,

$$\therefore w = \int_{-\infty}^t p \, dt = \int_{-\infty}^t i^2 R \, dt = \int_{-\infty}^t vi \, dt$$

Key Point: As i^2 term is always positive, the energy absorbed by the resistance is always positive.

If the voltage across resistance is constant V and the current through it is constant I then the energy for $t \geq 0$ is given by,

$$W = \int_0^t VI \, dt = VI t \text{ joules}$$

while, $P = VI = \frac{V^2}{R} = I^2 R \text{ watts}$

Factors Upon Which Resistance Depends

The resistance R of a conductor

(i) is directly proportional to its length *i.e.*

$$R \propto l$$

(ii) is inversely proportional to its area of X -section *i.e.*

$$R \propto \frac{1}{a}$$

(iii) depends upon the nature of material.

(iv) depends upon temperature.

From the first three points (leaving temperature for the time being), we have,

$$R \propto \frac{l}{a} \quad \text{or} \quad R = \rho \frac{l}{a}$$

where ρ (Greek letter 'Rho') is a constant and is known as *resistivity* or *specific resistance* of the material. Its value depends upon the nature of the material.

Specific Resistance or Resistivity

We have seen above that $R = \rho \frac{l}{a}$

If $l = 1 \text{ m}$, $a = 1 \text{ m}^2$, then, $R = \rho$

Hence **specific resistance** of a material is the resistance offered by 1 m length of wire of material having an area of cross-section of 1 m^2 [See Fig. 1.11 (i)].

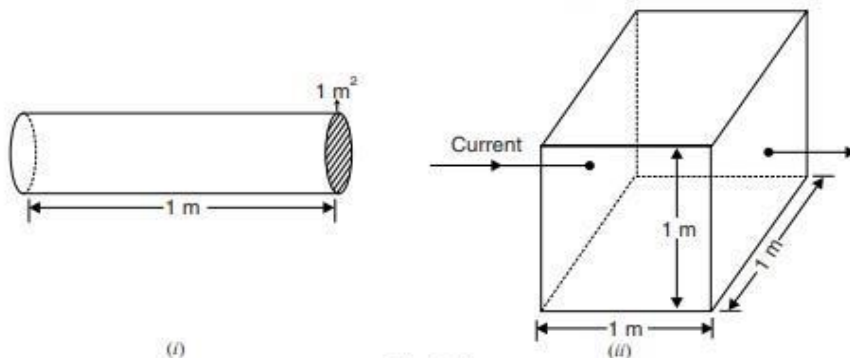


Fig. 1.11

Specific resistance can also be defined in another way. Take a cube of the material having each side 1 m. Considering any two opposite faces, the area of cross-section is 1 m^2 and length is 1 m [See Fig. 1.11 (ii)] *i.e.* $l = 1 \text{ m}$, $a = 1 \text{ m}^2$.

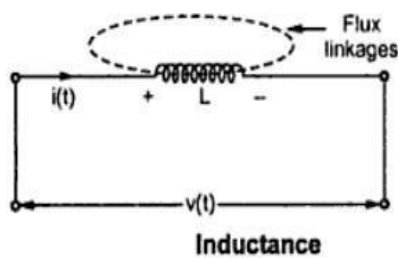
Hence **specific resistance** of a material may be defined as the resistance between the opposite faces of a metre cube of the material.

Unit of resistivity. We know $R = \frac{\rho l}{a}$ or $\rho = \frac{Ra}{l}$

Hence the unit of resistivity will depend upon the units of area of cross-section (a) and length (l).

2 Inductance

An inductance is the element in which energy is stored in the form of electromagnetic field. The inductance is denoted as 'L' and is measured in henries (H).



The Fig. 1.10 shows an inductance.

The time varying voltage $v(t)$ is the voltage across it. It carries a current $i(t)$ which is also time varying.

Key Point: For an inductance, the voltage across it is proportional to the rate of change of current passing through it.

$$\therefore v(t) \propto \frac{di(t)}{dt}$$

The constant of proportionality in the above equation is the inductance L.

$$\therefore v(t) = L \frac{di(t)}{dt}$$

If the voltage $v(t)$ is known across an inductor then the current is given by,

$$\therefore i(t) = \frac{1}{L} \int_{-\infty}^t v(t) dt$$

If the inductance has N turns and the flux ϕ produced by the current $i(t)$ entirely links with the coil of N turns then according to Faraday's law,

$$v(t) = N \frac{d\phi}{dt}$$

The total flux linkages $N\phi$ are thus proportional to the current through the coil.

$$\therefore N\phi = Li$$

$$\therefore L = \frac{N\phi}{i}$$

The power in the inductor is given by,

$$p(t) = vi = Li(t) \frac{di(t)}{dt}$$

The energy stored in the inductor in the form of an electromagnetic field is,

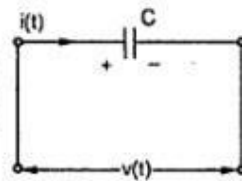
$$w = \int p(t) dt = \int Li(t) \frac{di(t)}{dt} dt$$

$$w = \int Li(t) di(t) = L \frac{i^2(t)}{2}$$

$$\therefore w = \frac{1}{2} Li^2(t) \text{ joules}$$

3 Capacitance

An element in which energy is stored in the form of an electrostatic field is known as capacitance. It is made up of two conducting plates separated by a dielectric material. It is denoted as 'C' and is measured in farads (F).



Capacitor

The Fig. shows a capacitor. The voltage across it is time varying $v(t)$ and current through it is also time varying $i(t)$.

Key Point: For a capacitor, the current through it is proportional to the rate of change of voltage across it.

$$i(t) \propto \frac{dv(t)}{dt}$$

The constant of proportionality is the capacitor C.

$$\therefore \quad i(t) = C \frac{dv(t)}{dt}$$

While the ratio of the charge stored to the voltage across the capacitor is known as the capacitance C.

$$\therefore \quad C = \frac{q}{v}$$

The voltage across the capacitor is given by,

$$v(t) = \frac{1}{C} \int_{-\infty}^t i(t) dt$$

The power in the capacitor is given by,

$$p(t) = v i = C v(t) \frac{dv(t)}{dt}$$

The energy stored in the capacitor is given by,

$$w = \int p(t) dt = \int C v(t) \frac{dv(t)}{dt} dt$$

$$w = \int C v(t) dv(t) = C \frac{v^2(t)}{2}$$

$$\therefore \quad w = \frac{1}{2} C v^2(t) \text{ joules}$$

Voltage Current Relationships for Passive Elements

The voltage current relationships for the passive elements resistance (R), inductance (L) and capacitor (C) are given in the Table 1.3.

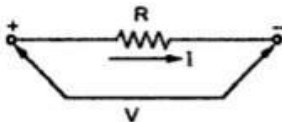
Element	Basic relation	Voltage across, if current known	Current through, if voltage known	Energy
R	$R = \frac{v}{i}$	$v_R(t) = R i_R(t)$	$i_R(t) = \frac{1}{R} v_R(t)$	$w = \int_{-\infty}^t i_R(t) v_R(t) dt$
L	$L = \frac{N\phi}{i}$	$v_L(t) = L \frac{di_L(t)}{dt}$	$i_L(t) = \frac{1}{L} \int_{-\infty}^t v_L(t) dt$	$w = \frac{1}{2} L i^2(t)$
C	$C = \frac{q}{v}$	$v_C(t) = \frac{1}{C} \int_{-\infty}^t i_C(t) dt$	$i_C(t) = C \frac{dv_C(t)}{dt}$	$w = \frac{1}{2} C v^2(t)$

Ohm's Law

This law gives relationship between the potential difference (V), the current (I) and the resistance (R) of a d.c. circuit. Dr. Ohm in 1827 discovered a law called Ohm's Law. It states,

Ohm's Law : *The current flowing through the electric circuit is directly proportional to the potential difference across the circuit and inversely proportional to the resistance of the circuit, provided the temperature remains constant.*

Mathematically, $I \propto \frac{V}{R}$



Where I is the current flowing in amperes, the V is the voltage applied and R is the resistance of the conductor, as shown in the Fig. 1.13.

$$\text{Now } I = \frac{V}{R}$$

Ohm's law

The unit of potential difference is defined in such a way that the constant of proportionality is unity.

Ohm's law is,

$I = \frac{V}{R}$	amperes
$V = I R$	volts
$\frac{V}{I} = \text{constant} = R$	ohms

The Ohm's law can be defined as,

The ratio of potential difference (V) between any two points of a conductor to the current (I) flowing between them is constant, provided that the temperature of the conductor remains constant.

Key Point: *Ohm's law can be applied either to the entire circuit or to the part of a circuit. If it is applied to entire circuit, the voltage across the entire circuit and resistance of the entire circuit should be taken into account. If the Ohm's law is applied to the part of a circuit, then the resistance of that part and potential across that part should be used.*

Limitations of Ohm's Law

The limitations of the Ohm's law are,

- 1) It is not applicable to the non linear devices such as diodes, zener diodes, voltage regulators etc.
- 2) It does not hold good for non-metallic conductors such as silicon carbide. The law for such conductors is given by,

$$V = k I^m \quad \text{where } k, m \text{ are constants.}$$

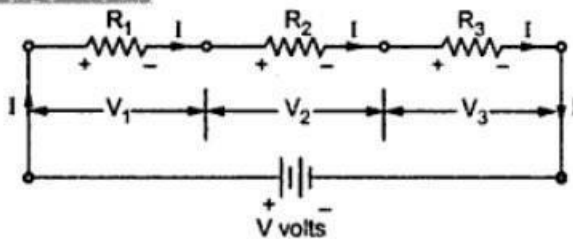
Series and Parallel Connections of Resistors on DC Excitation

D.C. Series Circuit

The d.c. circuit in which resistances are connected end to end so that there is only one path for current to flow is called a d.c. series circuit.

Resistors in Series

Current same
voltage division



A series circuit

Consider the resistances shown in the Fig.

The resistance R_1 , R_2 and R_3 are said to be in series. The combination is connected across a source of voltage V volts. Naturally the current flowing through all of them is same indicated as I amperes. e.g. the chain of small lights, used for the decoration purposes is good example of series combination.

Now let us study the voltage distribution.

Let V_1 , V_2 and V_3 be the voltages across the terminals of resistances R_1 , R_2 and R_3 respectively

Then,
$$V = V_1 + V_2 + V_3$$

Now according to Ohm's law,
$$V_1 = I R_1, V_2 = I R_2, V_3 = I R_3$$

Current through all of them is same i.e. I

\therefore
$$V = I R_1 + I R_2 + I R_3 = I(R_1 + R_2 + R_3)$$

Applying Ohm's law to overall circuit,

$$V = I R_{eq}$$

where R_{eq} = Equivalent resistance of the circuit. By comparison of two equations,

$$R_{eq} = R_1 + R_2 + R_3$$

i.e. total or equivalent resistance of the series circuit is arithmetic sum of the resistances connected in series.

For n resistances in series,

$$R = R_1 + R_2 + R_3 + \dots + R_n$$

Characteristics of Series Circuits

- 1) The same current flows through each resistance.
- 2) The supply voltage V is the sum of the individual voltage drops across the resistances.

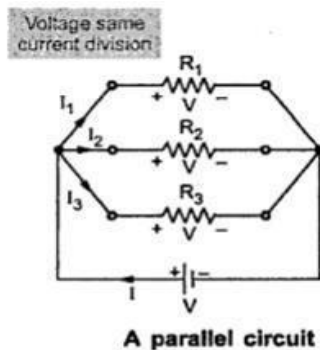
$$V = V_1 + V_2 + \dots + V_n$$

- 3) The equivalent resistance is equal to the sum of the individual resistances.
- 4) The equivalent resistance is the largest of all the individual resistances.

i.e.
$$R > R_1, R > R_2, \dots, R > R_n$$

D.C. Parallel Circuit

When one end of each resistance is joined to a common point and the other end of each resistance is joined to another common point so that there are as many paths for current flow as the number of resistances, it is called a parallel circuit.



In parallel circuit current passing through each resistance is different. Let total current drawn is say 'I' as shown. There are 3 paths for this current, one through R_1 , second through R_2 and third through R_3 . Depending upon the values of R_1 , R_2 and R_3 the appropriate fraction of total current passes through them. These individual currents are shown as I_1 , I_2 and I_3 . While the voltage across the two ends of each resistances R_1 , R_2 and R_3 is the same and equals the supply voltage V .

Now let us study current distribution. Apply Ohm's law to each resistance.

$$V = I_1 R_1, \quad V = I_2 R_2, \quad V = I_3 R_3$$

$$I_1 = \frac{V}{R_1}, \quad I_2 = \frac{V}{R_2}, \quad I_3 = \frac{V}{R_3}$$

$$I = I_1 + I_2 + I_3 = \frac{V}{R_1} + \frac{V}{R_2} + \frac{V}{R_3}$$

$$= V \left[\frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3} \right] \quad \dots (1)$$

For overall circuit if Ohm's law is applied,

$$V = I R_{eq}$$

and
$$I = \frac{V}{R_{eq}} \quad \dots (2)$$

where R_{eq} = Total or equivalent resistance of the circuit.

Comparing the two equations,

$$\frac{1}{R_{eq}} = \frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3}$$

where R is the equivalent resistance of the parallel combination.

In general if 'n' resistances are connected in parallel,

$$\frac{1}{R} = \frac{1}{R_1} + \frac{1}{R_2} + \frac{1}{R_3} + \dots + \frac{1}{R_n}$$

Characteristics of Parallel Circuits

- 1) The same potential difference gets across all the resistances in parallel.
- 2) The total current gets divided into the number of paths equal to the number of resistances in parallel. The total current is always sum of all the individual currents.

$$I = I_1 + I_2 + I_3 + \dots + I_n$$

- 3) The reciprocal of the equivalent resistance of a parallel circuit is equal to the sum of the reciprocal of the individual resistances.
- 4) The equivalent resistance is the smallest of all the resistances.

$$R < R_1, \quad R < R_2, \dots, R < R_n$$

- 5) The equivalent conductance is the arithmetic addition of the individual conductances.

Key Point : The equivalent resistance is smaller than the smallest of all the resistances connected in parallel.

Advantages of Parallel Circuits

The most useful property of a parallel circuit is the fact that potential difference has the same value between the terminals of each branch of parallel circuit. This feature of the parallel circuit offers the following advantages :

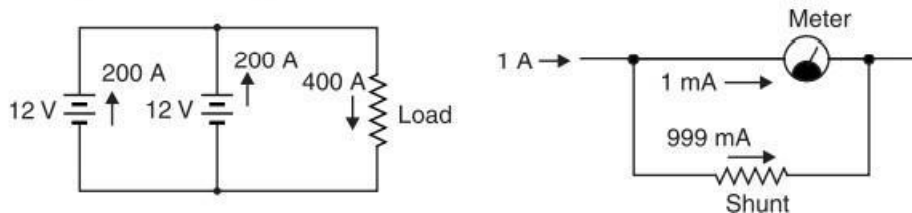
- (i) The appliances rated for the same voltage but different powers can be connected in parallel without disturbing each other's performance. Thus a 230 V, 230 W TV receiver can be operated independently in parallel with a 230 V, 40 W lamp.
- (ii) If a break occurs in any one of the branch circuits, it will have no effect on other branch circuits.

Due to above advantages, electrical appliances in homes are connected in parallel. We can switch on or off any light or appliance without affecting other lights or appliances.

Applications of Parallel Circuits

Parallel circuits find many applications in electrical and electronic circuits. We shall give two applications by way of illustration.

- (i) Identical voltage sources may be connected in parallel to provide a greater current capacity. Fig. 2.8 shows two 12 V automobile storage batteries in parallel. If the starter motor draws 400 A at starting, then each battery will supply half the current *i.e.* 200 A. A single battery might not be able to provide a load current of 400 A. Another benefit is that two batteries in parallel will supply a given load current for twice the time when compared to a single battery before discharge is reached.



- (ii) Fig. 2.9 shows another application for parallel connection. A low resistor, called a *shunt*, is connected in parallel with an ammeter to increase the current range of the meter. If shunt is not used, the ammeter is able to measure currents up to 1 mA. However, the use of shunt permits to measure currents up to 1 A. Thus shunt increases the range of the ammeter.

Network Terminology

While discussing network theorems and techniques, one often comes across the following terms:

- (i) **Linear circuit.** A linear circuit is one whose parameters (*e.g.* resistances) are constant *i.e.* they do not change with current or voltage.
- (ii) **Non-linear circuit.** A non-linear circuit is one whose parameters (*e.g.* resistances) change with voltage or current.
- (iii) **Bilateral circuit.** A bilateral circuit is one whose properties are the same in either direction. For example, transmission line is a bilateral circuit because it can be made to perform its function equally well in either direction.
- (iv) **Active element.** An active element is one which supplies electrical energy to the circuit. Thus in Fig. 3.1, E_1 and E_2 are the active elements because they supply energy to the circuit.
- (v) **Passive element.** A passive element is one which receives electrical energy and then either converts it into heat (resistance) or stores in an electric field (capacitance) or magnetic field (inductance). In Fig. 3.1, there are three passive elements, namely R_1 , R_2 and R_3 . These passive elements (*i.e.* resistances in this case) receive energy from the active elements (*i.e.* E_1 and E_2) and convert it into heat.
- (vi) **Node.** A node of a network is an equipotential surface at which *two or more* circuit elements are joined. Thus in Fig. 3.1, circuit elements R_1 and E_1 are joined at A and hence A is the node. Similarly, B , C and D are nodes.
- (vii) **Junction.** A junction is that point in a network where *three or more* circuit elements are joined. In Fig. 3.1, there are only two junction points *viz.* B and D . That B is a junction is clear from the fact that three circuit elements R_1 , R_2 and R_3 are joined at it. Similarly, point D is a junction because it joins three circuit elements R_2 , E_1 and E_2 .
- (viii) **Branch.** A branch is that part of a network which lies between two junction points. Thus referring to Fig. 3.1, there are a total of three branches *viz.* BAD , BCD and BD . The branch BAD consists of R_1 and E_1 ; the branch BCD consists of R_3 and E_2 and branch BD merely consists of R_2 .
- (ix) **Loop.** A loop is any closed path of a network. Thus in Fig. 3.1, $ABDA$, $BCDB$ and $ABCD$ are the loops.
- (x) **Mesh.** A mesh is the most elementary form of a loop and cannot be further divided into other loops. In Fig. 3.1, both loops $ABDA$ and $BCDB$ qualify as meshes because they cannot be further divided into other loops. However, the loop $ABCD$ cannot be called a mesh because it encloses two loops $ABDA$ and $BCDB$.
- (xi) **Network and circuit.** Strictly speaking, the term network is used for a circuit containing passive elements only while the term circuit implies the presence of both active and passive elements. However, there is no hard and fast rule for making these distinctions and the terms “network” and “circuit” are often used interchangeably.
- (xii) **Parameters.** The various elements of an electric circuit like resistance (R), inductance (L) and capacitance (C) are called parameters of the circuit. These parameters may be lumped or distributed.
- (xiii) **Unilateral circuit.** A unilateral circuit is one whose properties change with the direction of its operation. For example, a diode rectifier circuit is a unilateral circuit. It is because a diode rectifier cannot perform rectification in both directions.
- (xiv) **Active and passive networks.** An active network is that which contains active elements as well as passive elements. On the other hand, a passive network is that which contains passive elements only.

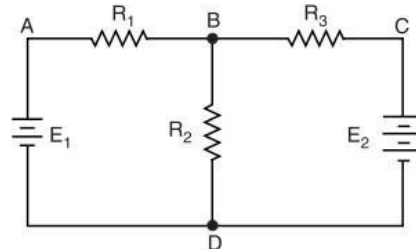


Fig. 3.1

Kirchhoff's Laws

Kirchhoff's Laws

Kirchhoff gave two laws to solve complex circuits, namely ;

1. Kirchhoff's Current Law (KCL)
2. Kirchhoff's Voltage Law (KVL)

1. KIRCHHOFF'S CURRENT LAW (KCL)

This law relates to the currents at the *junctions of an electric circuit and may be stated as under :

The algebraic sum of the currents meeting at a junction in an electrical circuit is zero.

An algebraic sum is one in which the sign of the quantity is taken into account. For example, consider four conductors carrying currents I_1, I_2, I_3 and I_4 and meeting at point O as shown in Fig. 2.60.

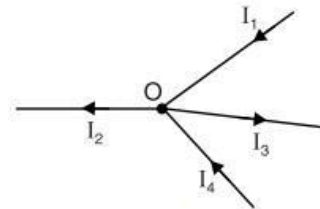


Fig. 2.60

* A junction is that point in an electrical circuit where *three or more* circuit elements meet.

If we take the signs of currents flowing towards point O as positive, then currents flowing away from point O will be assigned negative sign. Thus, applying Kirchhoff's current law to the junction O in Fig. 2.60, we have,

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

$$\text{or} \quad I_1 + I_4 = I_2 + I_3$$

i.e., Sum of incoming currents = Sum of outgoing currents

Hence, Kirchhoff's current law may also be stated as under :

The sum of currents flowing towards any junction in an electrical circuit is equal to the sum of currents flowing away from that junction. Kirchhoff's current law is also called junction rule.

Kirchhoff's current law is true because electric current is merely the flow of free electrons and they cannot accumulate at any point in the circuit. This is in accordance with the law of conservation of charge. Hence, Kirchhoff's current law is based on the law of conservation of charge.

2. KIRCHHOFF'S VOLTAGE LAW (KVL)

This law relates to *e.m.fs* and voltage drops in a closed circuit or loop and may be stated as under :

In any closed electrical circuit or mesh, the algebraic sum of all the electromotive forces (e.m.fs) and voltage drops in resistors is equal to zero, i.e.,

In any closed circuit or mesh,

$$\text{Algebraic sum of e.m.fs} + \text{Algebraic sum of voltage drops} = 0$$

The validity of Kirchhoff's voltage law can be easily established by referring to the closed loop $ABCD$ shown in Fig. 2.61. If we start from any point (say point A) in this closed circuit and go back to this point (*i.e.*, point A) after going around the circuit, then there is no increase or decrease in potential. This means that algebraic sum of the *e.m.fs* of all the sources (here only one *e.m.f.* source is considered) met on the way *plus* the algebraic sum of the voltage drops in the resistances must be zero. Kirchhoff's voltage law is based on the law of *conservation of energy, *i.e.*, net change in the energy of a charge after completing the closed path is zero.

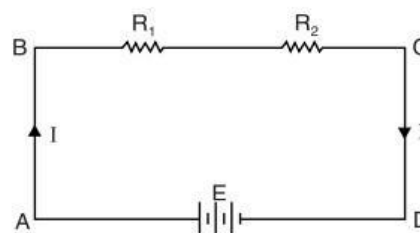


Fig. 2.61

Sign Convention

While applying Kirchhoff's voltage law to a closed circuit, algebraic sums are considered. Therefore, it is very important to assign proper signs to *e.m.fs* and voltage drops in the closed circuit. The following convention may be followed :

A rise in potential should be considered positive and fall in potential should be considered negative.

(i) Thus if we go from the positive terminal of the battery to the negative terminal, there is fall of potential and the *e.m.f.* should be assigned negative sign. Thus in Fig. 2.62 (i), as we go from *A* to *B*, there is a fall in potential and the *e.m.f.* of the cell will be assigned negative

sign. On the other hand, if we go from the negative terminal to the positive terminal of the battery or source, there is a rise in potential and the *e.m.f.* should be

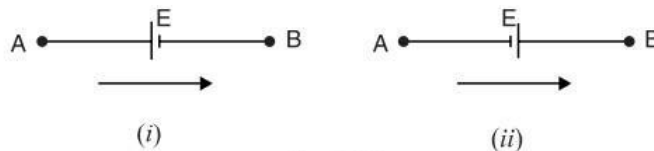


Fig. 2.62

assigned positive sign. Thus in Fig. 2.62 (ii) as we go from *A* to *B*, there is a rise in potential and the *e.m.f.* of the cell will be assigned positive sign. *It may be noted that the sign of e.m.f. is independent of the direction of current through the branch under consideration.*

(ii) When current flows through a resistor, there is a voltage drop across it. If we go through the resistor in the same direction as the current, there is a fall in potential because current flows from higher potential to lower potential. Hence this voltage drop should be assigned negative sign. In Fig. 2.63 (i), as we go from *A* to *B*, there is a fall in potential and the voltage drop across the resistor will be assigned negative sign.

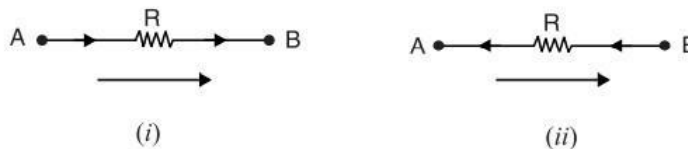


Fig. 2.63

On the other hand, if we go through the resistor against the current flow, there is a rise in potential and the voltage drop should be given positive sign. Thus referring to Fig. 2.63 (ii), as we go from *A* to *B*, there is a rise in potential and this voltage drop will be given positive sign. *It may be noted that sign of voltage drop depends on the direction of current and is independent of the polarity of the e.m.f. of source in the circuit under consideration.*

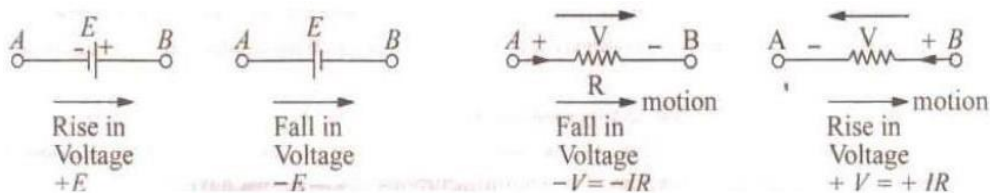
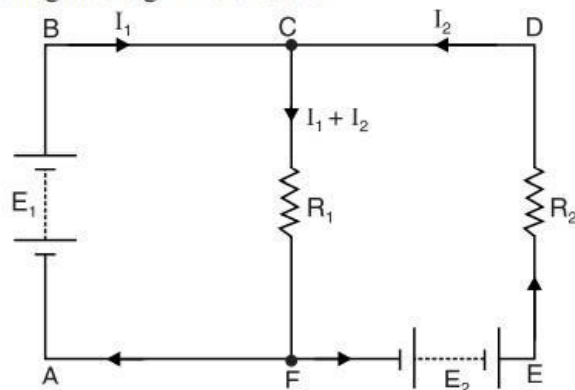


Illustration of Kirchhoff's Laws

Kirchhoff's Laws can be beautifully explained by referring to Fig. Mark the directions of currents as indicated. The direction in which currents are assumed to flow is unimportant, since if wrong direction is chosen, it will be indicated by a negative sign in the result.

(i) The magnitude of current in any branch of the circuit can be found by applying Kirchhoff's current law. Thus at junction C in Fig. 2.64, the incoming currents to the junction are I_1 and I_2 . Obviously, the current in branch CF will be $I_1 + I_2$.

(ii) There are three closed circuits in Fig 2.64 viz. $ABCFA$, $CDEFC$ and $ABCDEFA$. Kirchhoff's voltage law can be applied to these closed circuits to get the desired equations.



Loop ABCFA. In this loop, *e.m.f.* E_1 will be given *positive* sign. It is because as we consider the loop in the order $ABCFA$, we go from -ve terminal to the positive terminal of the battery in the branch AB and hence there is a rise in potential. The voltage drop in branch CF is $(I_1 + I_2) R_1$ and shall bear *negative* sign. It is because as we consider the loop in the order $ABCFA$, we go with current in branch CF and there is a fall in potential. Applying Kirchhoff's voltage law to the loop $ABCFA$,

$$-(I_1 + I_2) R_1 + E_1 = 0$$

or $E_1 = (I_1 + I_2) R_1$... (i)

Loop CDEFC. As we go around the loop in the order $CDEFC$, drop $I_2 R_2$ is *positive*, *e.m.f.* E_2 is *negative* and drop $(I_1 + I_2) R_1$ is *positive*. Therefore, applying Kirchhoff's voltage law to this loop, we get,

$$I_2 R_2 + (I_1 + I_2) R_1 - E_2 = 0$$

or $I_2 R_2 + (I_1 + I_2) R_1 = E_2$... (ii)

Since E_1 , E_2 , R_1 and R_2 are known, we can find the values of I_1 and I_2 from the above two equations. Hence currents in all branches can be determined.

Method to Solve Circuits by Kirchhoff's Laws

- (i) Assume unknown currents in the given circuit and show their direction by arrows.
- (ii) Choose any closed circuit and find the algebraic sum of voltage drops *plus* the algebraic sum of *e.m.fs* in that loop.
- (iii) Put the algebraic sum of voltage drops plus the algebraic sum of *e.m.fs* equal to zero.
- (iv) Write equations for as many closed circuits as the number of unknown quantities. Solve equations to find unknown currents.
- (v) If the value of the assumed current comes out to be negative, it means that actual direction of current is opposite to that of assumed direction.

Note. It may be noted that Kirchhoff's laws are also applicable to a.c. circuits. The only thing to be done is that \mathbf{I} , \mathbf{V} and \mathbf{Z} are substituted for I , V and R . Here \mathbf{I} , \mathbf{V} and \mathbf{Z} are phasor quantities.

Super positionTheorem

Superposition Theorem

Superposition is a general principle that allows us to determine the effect of several energy sources (voltage and current sources) acting simultaneously in a circuit by considering the effect of each source acting alone, and then combining (superposing) these effects. This theorem as applied to d.c. circuits may be stated as under :

In a linear, bilateral d.c. network containing more than one energy source, the resultant potential difference across or current through any element is equal to the algebraic sum of potential differences or currents for that element produced by each source acting alone with all other independent ideal voltage sources replaced by short circuits and all other independent ideal current sources replaced by open circuits (non-ideal sources are replaced by their internal resistances).

Procedure. The procedure for using this theorem to solve d.c. networks is as under :

- (i) Select one source in the circuit and replace all other ideal voltage sources by short circuits and ideal current sources by open circuits.
- (ii) Determine the voltage across or current through the desired element/branch due to single source selected in step (i).
- (iii) Repeat the above two steps for each of the remaining sources.
- (iv) Algebraically add all the voltages across or currents through the element/branch under consideration. The sum is the actual voltage across or current through that element/branch when all the sources are acting simultaneously.

Note. This theorem is called *superposition* because we superpose or algebraically add the components (currents or voltages) due to each independent source acting alone to obtain the total current in or voltage across a circuit element.

Example

Using superposition theorem, find the current through the 40 Ω resistor in the circuit shown in Fig. 3.45 (i). All resistances are in ohms.

Solution. In Fig. 3.45 (ii), 10V battery is replaced by a short so that 50V battery is acting alone. It can be seen that right-hand 5 Ω resistance is in parallel with 40 Ω resistance and their combined resistance = $5\ \Omega \parallel 40\ \Omega = 4.44\ \Omega$ as shown in Fig. 3.45 (iii). The 4.44 Ω resistance is in series with left-hand 5 Ω resistance giving total resistance of $(5 + 4.44) = 9.44\ \Omega$ to this path. As can be seen from Fig. 3.45 (iii), there are two parallel branches of resistances 20 Ω and 9.44 Ω across the 50 V battery. Therefore, current through 9.44 Ω branch is $I = 50/9.44 = 5.296\ \text{A}$. Thus in Fig. 3.45 (ii), the current $I (= 5.296\ \text{A})$ at point A divides between 5 Ω resistance and 40 Ω resistance. By current-divider rule, current I_1 in 40 Ω resistance is

$$I_1 = I \times \frac{5}{5 + 40} = 5.296 \times \frac{5}{45} = 0.589\ \text{A downward}$$

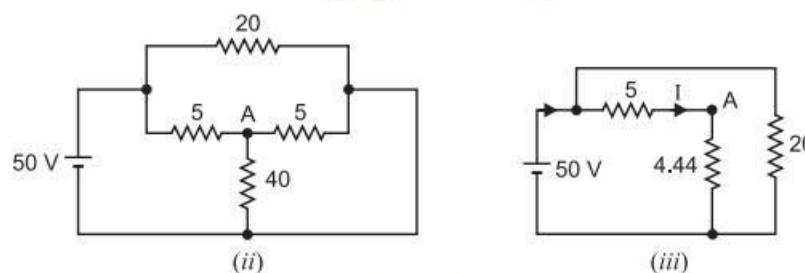
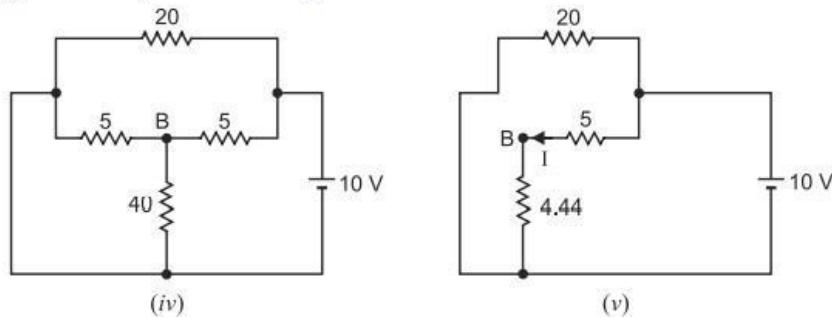


Fig. 3.45

In Fig. 3.45 (iv), the 50 V battery is replaced by a short so that 10 V battery is acting alone. Again, there are two parallel branches of resistances 20 Ω and 9.44 Ω across the 10V battery [See

Fig. 3.45 (v)]. Therefore, current through 9.44 Ω branch is $I = 10/9.44 = 1.059\ \text{A}$.

In Fig. 3.45 (iv), the 50 V battery is replaced by a short so that 10 V battery is acting alone. Again, there are two parallel branches of resistances $20\ \Omega$ and $9.44\ \Omega$ across the 10V battery [See Fig. 3.45 (v)]. Therefore, current through $9.44\ \Omega$ branch is $I = 10/9.44 = 1.059\ \text{A}$.



Thus in Fig. 3.45 (iv), the current $I (= 1.059\ \text{A})$ at point B divides between $5\ \Omega$ resistance and $40\ \Omega$ resistance. By current-divider rule, current in $40\ \Omega$ resistance is

$$I_2 = 1.059 \times \frac{5}{5 + 40} = 0.118\ \text{A downward}$$

∴ By superposition theorem, the total current in $40\ \Omega$

$$= I_1 + I_2 = 0.589 + 0.118 = \mathbf{0.707\ \text{A downward}}$$

Representation Of Sinusoidal Waveforms

Important A.C. Terminology

An alternating voltage or current changes continuously in magnitude and alternates in direction at regular intervals of time. It rises from zero to maximum positive value, falls to zero, increases to a maximum in the reverse direction and falls back to zero again (See Fig. 11.6). From this point on indefinitely, the voltage or current repeats the procedure. The important a.c. terminology is defined below :

(i) **Waveform.** The shape of the curve obtained by plotting the instantaneous values of voltage or current as ordinate against time as abscissa is called its *waveform* or *waveshape*. Fig. 11.6 shows the waveform of an alternating voltage varying sinusoidally.

(ii) **Instantaneous value.** The value of an alternating quantity at any instant is called instantaneous value. The instantaneous values of alternating voltage and current are represented by v and i respectively. As an example, the instantaneous values of voltage (See Fig. 11.6) at 0° , 90° and 270° are 0 , $+V_m$, $-V_m$ respectively.

(iii) **Cycle.** One complete set of positive and negative values of an alternating quantity is known as a cycle. Fig. 11.6 shows one cycle of an alternating voltage.

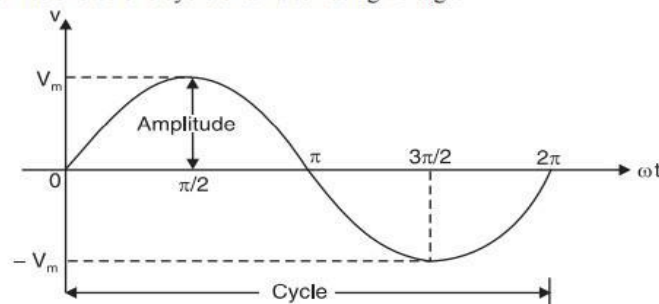


Fig. 11.6

A cycle can also be defined in terms of angular measure. One cycle corresponds to 360° electrical or 2π radians. The voltage or current generated in a conductor will span 360° electrical (or complete one cycle) when the conductor moves past successive north and south poles.

(iv) **Alternation.** One-half cycle of an alternating quantity is called an alternation. An alternation spans 180° electrical. Thus in Fig. 11.6, the positive or negative half of alternating voltage is the alternation.

(v) **Time period.** The time taken in seconds to complete one cycle of an alternating quantity is called its time period. It is generally represented by T .

(vi) **Frequency.** The number of cycles that occur in one second is called the frequency (f) of the alternating quantity. It is measured in cycles/sec (C/s) or Hertz (Hz). One Hertz is equal to 1C/s.

The frequency of power system is low; the most common being 50 C/s or 50 Hz. It means that alternating voltage or current completes 50 cycles in one second. The 50 Hz frequency is the most popular because it gives the best results when used for operating both lights and machinery.

(vii) **Amplitude.** The maximum value (positive or negative) attained by an alternating quantity is called its amplitude or peak value. The amplitude of an alternating voltage or current is designated by V_m (or E_m) or I_m .

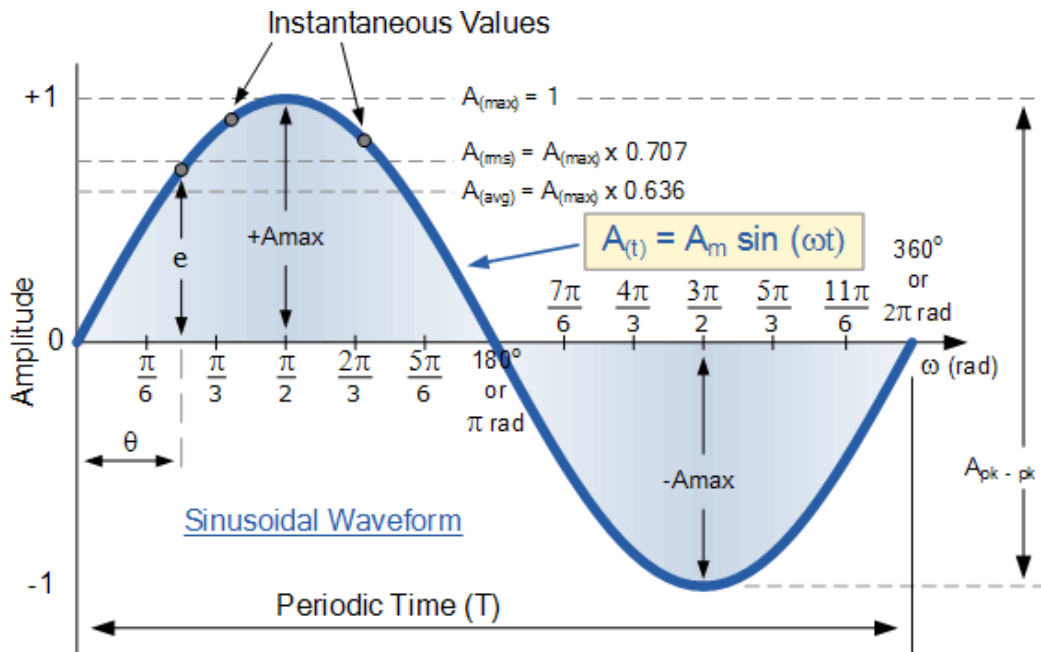


Fig:AC Waveform General Representation

Example An alternating current i is given by ;

$$i = 141.4 \sin 314 t$$

Find (i) the maximum value (ii) frequency (iii) time period and (iv) the instantaneous value when t is 3 ms.

Solution. Comparing the given equation of alternating current with the standard form $i = I_m \sin \omega t$, we have,

- (i) Maximum value, $I_m = 141.4 \text{ A}$
- (ii) Frequency, $f = \omega/2\pi = 314/2\pi = 50 \text{ Hz}$
- (iii) Time period, $T = 1/f = 1/50 = 0.02 \text{ s}$
- (iv) $i = 141.4 \sin 314 t$
When $t = 3 \text{ m s} = 3 \times 10^{-3} \text{ s}$,
 $i = 141.4 \sin 314 \times 3 \times 10^{-3} = 114.35 \text{ A}$

Peak and RMS Values

Values of Alternating Voltage and Current

In a d.c. system, the voltage and current are constant so that there is no problem of specifying their magnitudes. However, an alternating voltage or current varies from instant to instant. A natural question arises how to express the magnitude of an alternating voltage or current. There are four ways of expressing it, namely ;

- (i) Peak value
- (ii) Average value or mean value
- (iii) R.M.S. value or effective value
- (iv) Peak-to-peak value

Although peak, average and peak-to-peak values may be important in some engineering applications, it is the r.m.s. or effective value which is used to express the magnitude of an alternating voltage or current.

Peak Value

It is the maximum value attained by an alternating quantity. The peak or maximum value of an alternating voltage or current is represented by V_m or I_m . The knowledge of peak value is important in case of testing materials. However, peak value is not used to specify the magnitude of alternating voltage or current. Instead, we generally use r.m.s. values to specify alternating voltages and currents.

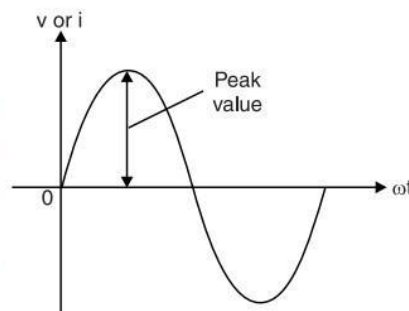


Fig. 11.14

Average Value

The average value of a waveform is the average of all its values over a period of time. In performing such a computation, we regard the area above the time axis as positive area and area below the time axis as negative area. The algebraic signs of the areas must be taken into account when computing the total (net) area. The time interval over which the net area is computed is the period T of the waveform.

$$\therefore \text{Average value} = \frac{\text{Total (net) area under curve for time } T}{\text{Time } T}$$

Average Value of Sinusoidal Current

The average value of alternating current (or voltage) over one cycle is zero. It is because the waveform is symmetrical about time axis and positive area exactly cancels the negative area. However, the average value over a half-cycle (positive or negative) is not zero. Therefore, average value of alternating current (or voltage) means half-cycle average value unless stated otherwise.

The half-cycle average value of a.c. is that value of steady current (d.c.) which would send the same amount of charge through a circuit for half the time period of a.c. as is sent by the a.c. through the same circuit in the same time. It is represented by I_{av} . This can be obtained by integrating the instantaneous value of current over one half cycle (i.e. area over half-cycle) and dividing the result by base length of half-cycle ($= \pi$).

The equation of an alternating current varying sinusoidally is given by ;

$$i = I_m \sin \theta$$

Consider an elementary strip of thickness $d\theta$ in the first half-cycle of current wave as shown in Fig. 11.15. Let i be the mid-ordinate of this strip.

$$\text{Area of strip} = i d\theta$$

$$\begin{aligned} \text{Area of half-cycle} &= \int_0^{\pi} i d\theta \\ &= \int_0^{\pi} I_m \sin \theta d\theta \\ &= I_m [-\cos \theta]_0^{\pi} = 2I_m \end{aligned}$$

$$\therefore \text{Average value, } I_{av} = \frac{\text{Area of half-cycle}}{\text{Base length of half-cycle}} = \frac{2I_m}{\pi}$$

$$\text{or } I_{av} = 0.637 I_m$$

Hence, the half-cycle average value of a.c. is 0.637 times the peak value of a.c.

For positive half-cycle, $I_{av} = +0.637 I_m$

For negative half-cycle, $I_{av} = -0.637 I_m$

Clearly, average value of a.c. over a complete cycle is zero. Similarly, it can be proved that for alternating voltage varying sinusoidally, $V_{av} = 0.637 V_m$.

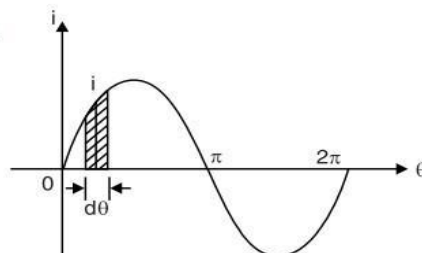


Fig. 11.15

R.M.S. or Effective Value

The average value cannot be used to specify a sinusoidal voltage or current. It is because its value over one-cycle is zero and cannot be used for power calculations. Therefore, we must search for a more suitable criterion to measure the effectiveness of an alternating current (or voltage). The obvious choice would be to measure it in terms of direct current that would do work (or produce heat) at the same average rate under similar conditions. This equivalent direct current is called the root-mean-square (r.m.s.) or effective value of alternating current.

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

For example, when we say that the r.m.s. or effective value of an alternating current is 5A, it means that the alternating current will do work (or produce heat) at the same rate as 5A direct current under similar conditions.

R.M.S. Value of Sinusoidal Current

The equation of the alternating current varying sinusoidally is given by ;

$$i = I_m \sin \theta$$

Consider an elementary strip of thickness $d\theta$ in first half-cycle of the squared current wave (shown dotted in Fig. 11.21). Let i^2 be the mid-ordinate of this strip.

$$\text{Area of strip} = i^2 d\theta$$

Area of half-cycle of the squared wave

$$\begin{aligned} &= \int_0^{\pi} i^2 d\theta \\ &= \int_0^{\pi} I_m^2 \sin^2 \theta d\theta \\ &= I_m^2 \int_0^{\pi} \sin^2 \theta d\theta = \frac{\pi I_m^2}{2} \end{aligned}$$

$$\therefore I_{r.m.s.} = \sqrt{\frac{\text{Area of half-cycle squared wave}}{\text{Half-cycle base}}}$$

$$= \sqrt{\frac{\pi I_m^2 / 2}{\pi}} = \frac{I_m}{\sqrt{2}} = 0.707 I_m$$

$$\therefore I_{r.m.s.} = 0.707 I_m$$

Similarly, it can be proved that for alternating voltage varying sinusoidally, $V_{r.m.s.} = 0.707 V_m$.

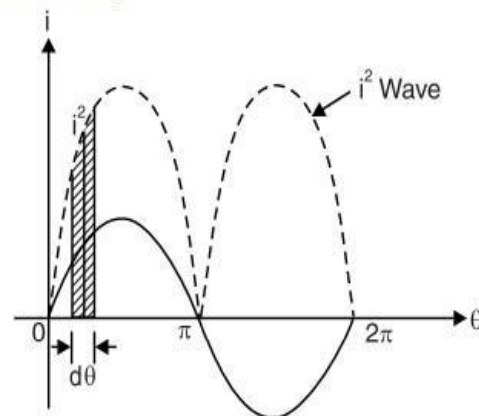


Fig. 11.21

Importance of R.M.S. Values

An alternating voltage or current is always specified in terms of r.m.s. values. For example, common household appliances are rated at 230 V a.c. This is an r.m.s. value. If some other method of measurement is used, it must be specifically stated. Lacking any information to the contrary, always assume that a.c. values are r.m.s. The following points will give the reader a clear concept about the r.m.s. values :

- (i) The domestic a.c. supply is 230 V, 50 Hz. It is the r.m.s. or effective value. It means that alternating voltage available has the same heating effect as 230 V d.c. The equation of this alternating voltage is given by ;

$$v = V_m \sin \omega t$$

$$= 230 \times \sqrt{2} \sin 2\pi \times 50 \times t \quad (V_m = \sqrt{2} V_{r.m.s.})$$

$$\therefore v = 230 \sqrt{2} \sin 314 t$$

- (ii) When we say that alternating current in a circuit is 5 A, we are specifying the r.m.s. value. It means that the alternating current flowing in the circuit has the same heating effect as 5 A d.c.

- (iii) A.C. ammeters and voltmeters record r.m.s. values of alternating current and voltage respectively.

Fig. 11.23 summarises the various ways to measure sinusoidal voltages and the conversion constants. The relationships apply for currents as well as voltages.

$$V_{p-p} = 2 \times V_m$$

$$V_{av} = 0.637 \times V_m$$

$$V_{r.m.s.} = 0.707 \times V_m$$

$$V_m = 1.414 \times V_{r.m.s.}$$

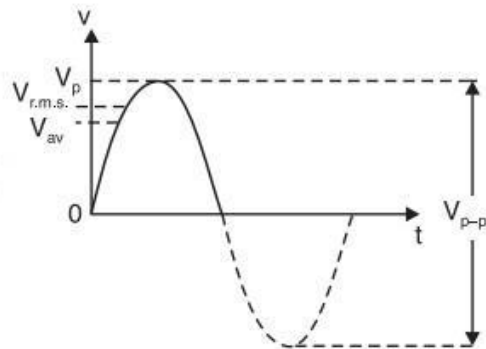


Fig. 11.23

It is important to note that above relationship between peak, average and r.m.s. quantities are applicable to **only pure sine waves**. In the case of other waveforms, these quantities are related by other (different) factors.

Note. R.M.S. value of an a.c. wave is always greater than the average value **except** in the case of rectangular and square waves when both are equal.

Form Factor and Peak Factor

There exists a definite relation among the peak value, average value and r.m.s. value of an alternating quantity. The relationship is expressed by two factors, namely ; form factor and peak factor.

(i) **Form factor.** The ratio of r.m.s. value to the average value of an alternating quantity is known as **form factor** i.e.

$$\text{Form factor} = \frac{\text{R.M.S. value}}{\text{Average value}}$$

The value of form factor depends upon the waveform of the alternating quantity. Its least value is 1 (e.g. for square wave, rectangular wave) and may be as high as 5 for other waveforms. The form factor for an alternating voltage or current varying sinusoidally is 1.11. *i.e.*

For a sinusoidal voltage or current,

$$\text{Form factor} = \frac{0.707 \times \text{Max. value}}{0.637 \times \text{Max. value}} = 1.11$$

The form factor gives a measure of the “peakiness” of the waveform. The peakier the wave, the greater is its form factor and *vice-versa*. For instance, a sine wave is peakier than a square wave. Hence the former has a greater form factor (1.11) than the latter. Similarly, a triangular wave is more peaky than a sine wave and has a form factor of 1.15. The form factor is useful in rectifier service.

(ii) Peak factor. *The ratio of maximum value to the r.m.s. value of an alternating quantity is known as peak factor i.e.*

$$\text{Peak factor} = \frac{\text{Max. value}}{\text{R.M.S. value}}$$

The value of peak factor also depends upon the waveform of the alternating quantity. For an alternating voltage or current varying sinusoidally, its value is 1.414 *i.e.*

For a sinusoidal voltage or current,

$$\text{Peak factor} = \frac{\text{Max. value}}{0.707 \times \text{Max. value}} = 1.414$$

The peak factor is of much greater importance because it indicates the maximum voltage being applied to the various parts of the apparatus. For instance, when an alternating voltage is applied across a cable or capacitor, the breakdown of insulation will depend upon the maximum voltage. The insulation must be able to withstand the maximum rather than the r.m.s. value of voltage.

Note. Peak factor is also called *crest factor* or *amplitude factor*.

Phasor Representation of Sinusoidal Quantities

Consider an alternating current represented by the equation $i = I_m \sin \omega t$. Take a line OP to represent to scale the maximum value I_m . Imagine the line OP (or *phasor*, as it is called) to be rotating in anticlockwise direction at an angular velocity ω rad/sec about the point O . Measuring the time from the instant when OP is horizontal, let OP rotate through an angle $\theta (= \omega t)$ in the anticlockwise direction. The projection of OP on the Y -axis is OM .

$$\begin{aligned} OM &= OP \sin \theta \\ &= I_m \sin \omega t \\ &= i, \text{ the value of current} \\ &\quad \text{at that instant} \end{aligned}$$

Hence the projection of the phasor OP on the Y -axis at any instant gives the value of current at that instant. Thus when $\theta = 90^\circ$, the projection on Y -axis is $OP (= I_m)$ itself. That the value of current at this instant (*i.e.* at θ or $\omega t = 90^\circ$) is I_m can be readily established if we put $\theta = 90^\circ$ in the current equation. If we plot the projections of the phasor on the Y -axis *versus* its angular position point-by-point, a sinusoidal alternating current wave is generated as shown in Fig. 11.47. Thus the phasor represents the sine wave for every instant of time.

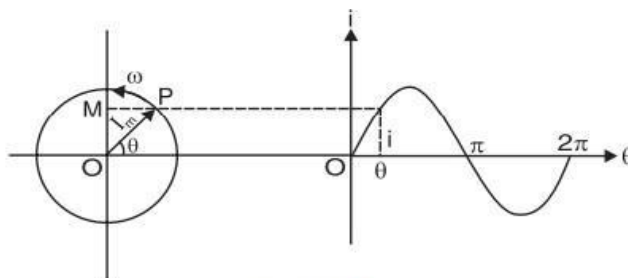


Fig. 11.47

The following points are worth noting :

- (i) The length of the phasor represents the maximum value and the angle with axis of reference (*i.e.*, X -axis) indicates the phase of the alternating quantity *i.e.* current in this case.
- (ii) The phasor representation enables us to quickly obtain the numerical values and, at the same time, have a picture before the eye of the events taking place in the circuit. Thus in the position of the phasor OP shown in Fig. 11.47, the instantaneous value is OM , the phase is θ and frequency is $\omega/2\pi$.
- (iii) A phasor diagram permits addition and subtraction of alternating voltages or currents with a fair degree of ease.

Phasor Diagram of Sine Waves of Same Frequency

Consider a sinusoidal voltage wave v and sinusoidal current wave i of the same frequency. Suppose the current lags behind the voltage by ϕ° . The two alternating quantities can be represented on the same phasor diagram because the phasors V_m and I_m [See Fig. 11.48 (i)] rotate at the same angular velocity ω and hence phase difference ϕ between them remains the same at all times. When each phasor completes one revolution, it generates the corresponding cycle [See Fig. 11.48 (ii)]. The equations of the two waves can be represented as :

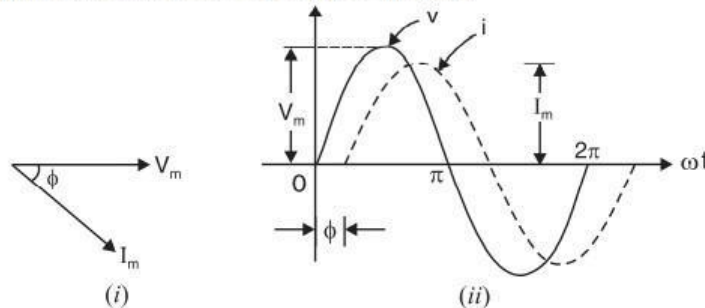


Fig. 11.48

$$v = V_m \sin \omega t$$

$$i = I_m \sin (\omega t - \phi)$$

The following points may be noted carefully :

- (i) The wave diagram and the phasor diagram convey the same information. However, it is more difficult to draw the waves than to sketch the phasor diagram.
- (ii) Since the two phasors have the same angular velocity (ω) and there is no relative motion between them, they can be displayed in a stationary diagram, the common angular rotation (ω) being disregarded.

Real Power, Reactive Power and Apparent Power

Apparent, True and Reactive Powers

Consider an inductive circuit in which circuit current I lags behind the applied voltage V by ϕ° . The phasor diagram of the circuit is shown in Fig. 12.6. The current I can be resolved into two rectangular components viz.

- (i) $I \cos \phi$ in phase with V .
- (ii) $I \sin \phi$; 90° out of phase with V .

1. Apparent power. The total power that appears to be transferred between the source and load is called **apparent power**. It is equal to the product of applied voltage (V) and circuit current (I) i.e.

$$\text{Apparent power, } S = V \times I = VI$$

It is measured in volt-amperes (VA).

Apparent power has two components viz true power and reactive power.

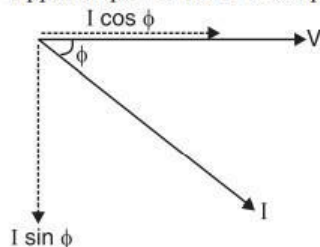


Fig. 12.6

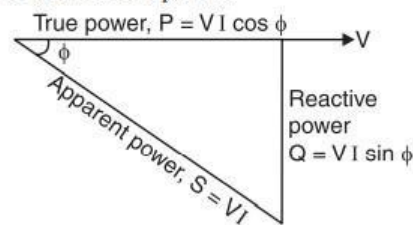


Fig. 12.7

2. True power. The power which is actually consumed in the circuit is called **true power** or **active power**. We know that power is consumed in resistance only since neither pure inductor (L) nor pure capacitor (C) consumes any active power. Now, current and voltage are in phase in a resistance. Therefore, current in phase with voltage produces true or active power. It is the useful component of apparent power.

The product of voltage (V) and component of total current in phase with voltage ($I \cos \phi$) is equal to **true power** i.e.

$$\begin{aligned}\text{True power, } P &= \text{Voltage} \times \text{Component of total current in phase with voltage} \\ &= V \times I \cos \phi\end{aligned}$$

$$\therefore P = VI \cos \phi$$

It is measured in watts (W). The component $I \cos \phi$ is called in-phase component or **wattful component** because it is this component of total current which contributes to true power (*i.e.* $VI \cos \phi$). It may be noted that it is the true power which is used for producing torque in motors and supply heat, light *etc.* It is used up in the circuit and cannot be recovered.

3. Reactive power. The component of apparent power which is neither consumed nor does any useful work in the circuit is called **reactive power**. The power consumed (or true power) in L and C is zero because all the power received from the source in one quarter-cycle is returned to the source in the next quarter-cycle. This circulating power is called *reactive power. Now, current and voltage in L or C are 90° out of phase. Therefore, current 90° out of phase with voltage contributes to reactive power.

*The product of voltage (V) and component of total current 90° out of phase with voltage ($I \sin \phi$) is equal to reactive power *i.e.**

$$\begin{aligned}\text{Reactive power, } Q &= \text{Voltage} \times \text{Component of total current } 90^\circ \text{ out of phase with voltage} \\ &= V \times I \sin \phi\end{aligned}$$

$$\therefore Q = VI \sin \phi$$

It is measured in volt-amperes reactive (VAR). The component $I \sin \phi$ is called the **reactive component** (or **wattless component**) and contributes to reactive power (*i.e.* $VI \sin \phi$). It does no useful work in the circuit and merely flows back and forth in both directions in the circuit. A wattmeter does not measure the reactive power.

Power triangle. If we multiply each of the current phasors in Fig. 12.6 by V , we get the power triangle shown in Fig. 12.7. This is a right-angled triangle and indicates the relation among apparent power, true power and reactive power. It reveals the following facts about the circuit :

- (i) Power factor, $\cos \phi = \frac{\text{True power}}{\text{Apparent power}} = \frac{VI \cos \phi}{VI}$
- (ii) $(\text{Apparent power})^2 = (\text{True power})^2 + (\text{Reactive power})^2$
or $S^2 = P^2 + Q^2$
- (iii) True power, $P = \text{Apparent power} \times \cos \phi = VI \cos \phi$
Reactive power, $Q = \text{Apparent power} \times \sin \phi = VI \sin \phi$

Illustration. Let us illustrate the power relations in an a.c. circuit with an example. Suppose a circuit draws a current of 10 A at a voltage of 200 V and its p.f. is 0.8 lagging. Then,

$$\text{Apparent power, } S = VI = 200 \times 10 = 2000 \text{ VA}$$

$$\text{True power, } P = VI \cos \phi = 200 \times 10 \times 0.8 = 1600 \text{ W}$$

$$\text{Reactive power, } Q = VI \sin \phi = 200 \times 10 \times 0.6 = 1200 \text{ VAR}$$

The circuit receives an apparent power of 2000 VA and is able to convert only 1600 watts into true power. The reactive power of 1200 VAR does no useful work, it merely flows into and out of the circuit periodically. In fact, reactive power is a liability on the source because the source has to supply the additional current (*i.e.* $I \sin \phi$) to provide for this power.

Power Factor

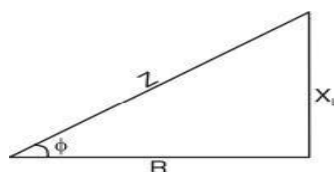


Fig. 12.5

Power Factor

The power factor (*i.e.* $\cos \phi$) of a circuit can be defined in one of the following ways :

- (i) Power factor = $\cos \phi$ = cosine of angle between V and I
- (ii) Power factor = $\frac{R}{Z} = \frac{\text{Resistance}}{\text{Impedance}}$ [See Fig. 12.5]
- (iii) Power factor = $\frac{VI \cos \phi}{VI} = \frac{\text{True power}}{\text{Apparent power}}$

For example, in a resistor, the current and voltage are in phase *i.e.* $\phi = 0^\circ$. Therefore, power factor of a pure resistive circuit is $\cos 0^\circ = 1$. Similarly, phase difference between voltage and current in a pure inductance or capacitance is 90° . Hence power factor of pure L or C is zero. This is the reason that power consumed by pure L or C is *zero. For a circuit having R , L and C in varying proportions, the value of power factor will lie between 0 and 1. It may be noted that power factor can never have a value greater than 1.

(a) It is a usual practice to attach the word ‘lagging’ or ‘leading’ with the numerical value of power factor to signify whether the current lags or leads the voltage. Thus if a circuit has a p.f. of 0.5 and the current lags the voltage, we generally write p.f. as 0.5 lagging.

(b) Sometimes power factor is expressed as a percentage. Thus 0.8 lagging power factor may be expressed as 80% lagging.

Significance of Power Factor

The apparent power drawn by a circuit has two components *viz.* (i) true power and (ii) reactive power. True power component should be as large as possible because it is this component which does useful work in the circuit. This is possible only if the reactive power component is small. As seen from the power triangle in Fig. 12.7, the smaller the phase angle ϕ (*i.e.* greater the p.f. $\cos \phi$), the smaller is the reactive power component. Thus when $\phi = 0^\circ$ (*i.e.* $\cos \phi = 1$), the reactive power component is zero and the true power is **equal to the apparent power. That means the whole of apparent power drawn by the circuit is being utilised by it. *Thus power factor of a circuit is a measure of its effectiveness in ***utilising the apparent power drawn by it.* The greater the power factor of a circuit, the greater is its ability to utilise the apparent power. Thus 0.5 p.f. (*i.e.* 50% p.f.) of a circuit means that it will utilise only 50% of the apparent power whereas 0.8 p.f. would mean 80% utilisation of apparent power. For this reason, we wish that the power factor of the circuit to be as near to 1 as possible.

Analysis Of Single-Phase Ac Series RL,RC And RLCCircuits

Introduction

An a.c. circuit differs from a d.c. circuit in many respects. First, in a d.c. circuit we consider resistances only whereas in an a.c. circuit, in addition to resistance (R), inductance (L) and capacitance (C) also play the part. The elements L and C offer opposition (*i.e.* X_L and X_C) to current flow in an a.c. circuit. Secondly, the magnitude of current in an a.c. circuit is affected by the supply frequency because $X_L (= 2\pi fL)$ and $X_C (= 1/2\pi fC)$ are frequency dependent. However, such a situation is not encountered in a d.c. circuit. Thirdly, in a d.c. circuit, voltages or currents can be added or subtracted arithmetically. But in an a.c. circuit, there is a phase difference of 90° between voltage across and current through L or C . This implies that for the addition or subtraction of alternating voltages or currents, phase difference has to be taken into account. All these features make the analysis of an a.c. circuit quite different from that of a d.c. circuit. In this chapter, we shall confine our attention to series a.c. circuits only.

R-L Series A.C. Circuit

This is the most general case met in practice as nearly all a.c. circuits contain both resistance and inductance. Fig. 12.1 (i) shows a pure resistance of R ohms connected in series with a coil of pure inductance L henry.

Let V = r.m.s. value of the applied voltage
 I = r.m.s. value of the circuit current
 $\therefore V_R = IR$ where V_R is in phase with I
 $V_L = IX_L$ where V_L leads I by 90°

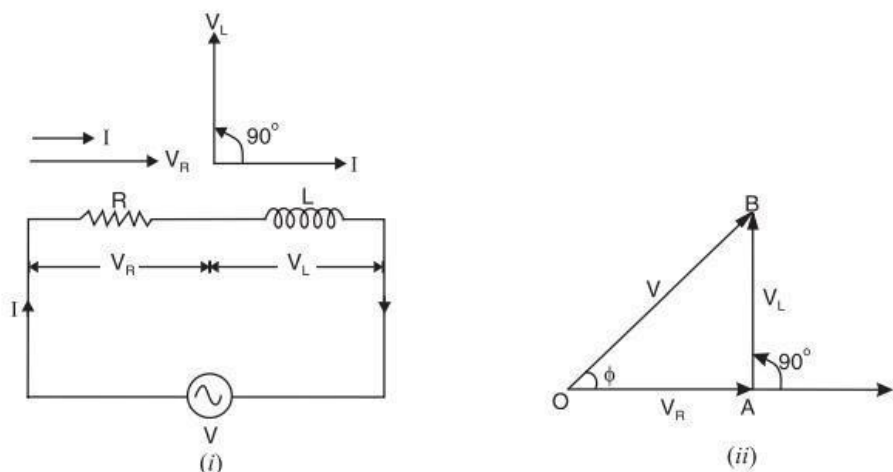


Fig. 12.1

Taking *current as the reference phasor, the phasor diagram of the circuit can be drawn as shown in Fig. 12.1 (ii). The voltage drop $V_R (= IR)$ is in phase with current and is represented in magnitude and direction by the phasor OA . The voltage drop $V_L (= IX_L)$ leads the current by 90° and is represented in magnitude and direction by the phasor AB . The applied voltage V is the phasor sum of these two drops i.e.

$$V = \sqrt{V_R^2 + V_L^2} = \sqrt{(IR)^2 + (IX_L)^2} = I\sqrt{R^2 + X_L^2}$$

$$\therefore I = \frac{V}{\sqrt{R^2 + X_L^2}}$$

The quantity $\sqrt{R^2 + X_L^2}$ offers opposition to current flow and is called **impedance** of the circuit. It is represented by Z and is measured in ohms (Ω).

$$\therefore I = \frac{V}{Z} \quad \text{where } Z = \sqrt{R^2 + X_L^2}$$

(i) Phase angle. It is clear from the phasor diagram that circuit current I lags behind the applied voltage V by ϕ° . This fact is also illustrated in the wave diagram shown in Fig. 12.2. The value of phase angle ϕ can be determined from the phasor diagram.

$$\tan \phi = \frac{V_L}{V_R} = \frac{IX_L}{IR} = \frac{X_L}{R}$$

Since X_L and R are known, ϕ can be calculated.

If the applied voltage is $v = V_m \sin \omega t$, then equation for the circuit current will be :

$$i = I_m \sin (\omega t - \phi) \quad \text{where } I_m = V_m / Z$$

We arrive at a very important conclusion that *in an inductive circuit, current lags behind the applied voltage*. The angle of lag (i.e. ϕ) is greater than 0° but less than 90° . It is determined by the ratio of inductive reactance to resistance ($\tan \phi = X_L/R$) in the circuit. The greater the value of this ratio, the greater will be the phase angle ϕ and *vice-versa*.

(ii) Impedance. The total opposition offered to the flow of alternating current by a circuit is called **impedance** Z of the circuit. In R - L series circuit,

$$\text{Impedance, } Z = \sqrt{R^2 + X_L^2} \quad \text{where } X_L = 2\pi fL$$

The magnitude of impedance in R - L series circuit depends upon the values of R , L and the supply frequency f .

(iii) Admittance (Y). Admittance of an a.c. circuit is the reciprocal of its impedance i.e.

$$\text{Admittance, } Y = \frac{1}{Z}$$

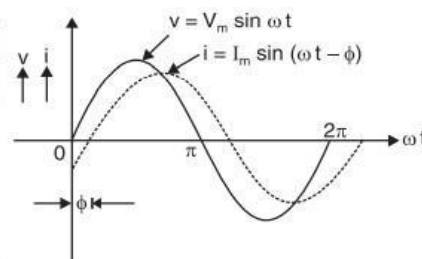


Fig. 12.2

Q-factor of a Coil

The ratio of the inductive reactance (X_L) of a coil to its resistance (R) at a given frequency is known as † Q -factor of the coil at that frequency *i.e.*,

$$Q\text{-factor} = \frac{X_L}{R} = \frac{\omega L}{R}$$

Also,
$$Q\text{-factor} = 2\pi \times \frac{\text{maximum energy stored}}{\text{energy dissipated per cycle}}$$

The Q -factor is used to describe the quality or effectiveness of a coil. A coil is usually designed to have high value of L compared to its resistance R . The greater the value of Q -factor of a coil, the greater is its inductance (L) as compared to its resistance (R). Many of the equations to be developed in a.c. circuit analysis can be simplified by the substitution of Q for the ratio X_L/R .

Example 12.1. A coil having a resistance of 7Ω and an inductance of 31.8 mH is connected to 230 V , 50 Hz supply. Calculate (i) the circuit current (ii) phase angle (iii) power factor (iv) power consumed and (v) voltage drop across resistor and inductor.

Solution. (i) Inductive reactance, $X_L = 2\pi fL = 2\pi \times 50 \times 31.8 \times 10^{-3} = 10 \Omega$

$$\text{Coil impedance, } Z = \sqrt{R^2 + X_L^2} = \sqrt{7^2 + 10^2} = 12.2 \Omega$$

$$\therefore \text{Circuit current, } I = V/Z = 230/12.2 = \mathbf{18.85 \text{ A}}$$

$$\text{(ii) } \tan \phi = X_L/R = 10/7$$

$$\therefore \text{Phase angle, } \phi = \tan^{-1}(10/7) = \mathbf{55^\circ \text{ lag}}$$

$$\text{(iii) Power factor} = \cos \phi = \cos 55^\circ = \mathbf{0.573 \text{ lag}}$$

$$\text{(iv) Power consumed, } P = VI \cos \phi = 230 \times 18.85 \times 0.573 = \mathbf{2484.24 \text{ W}}$$

$$\text{(v) Voltage drop across } R = IR = 18.85 \times 7 = \mathbf{131.95 \text{ V}}$$

$$\text{Voltage drop across } L = IX_L = 18.85 \times 10 = \mathbf{188.5 \text{ V}}$$

Example 12.2. An inductor coil is connected to a supply of 250 V at 50 Hz and takes a current of 5 A . The coil dissipates 750 W . Calculate (i) power factor (ii) resistance of coil and (iii) inductance of coil.

Solution. (i) Power consumed, $P = VI \cos \phi$

$$\therefore \text{Power factor, } \cos \phi = \frac{P}{VI} = \frac{750}{250 \times 5} = \mathbf{0.6 \text{ lag}}$$

$$\text{(ii) Impedance of coil, } Z = V/I = 250/5 = 50 \Omega$$

$$\text{Resistance of coil, } R = Z \cos \phi = 50 \times 0.6 = \mathbf{30 \Omega}$$

$$\text{(iii) Reactance of coil, } X_L = \sqrt{Z^2 - R^2} = \sqrt{(50)^2 - (30)^2} = 40 \Omega$$

$$\therefore \text{Inductance of coil, } L = \frac{X_L}{2\pi f} = \frac{40}{2\pi \times 50} = \mathbf{0.127 \text{ H}}$$

Example 12.3. A pure inductance of 318 mH is connected in series with a pure resistance of 75Ω . The circuit is supplied from 50 Hz source and the voltage across 75Ω resistor is found to be 150 V . Calculate the supply voltage and the phase angle.

Solution. The circuit diagram and the phasor diagram are shown in Fig. 12.8.

$$\text{Circuit current, } I = V_R/R = 150/75 = 2 \text{ A}$$

$$\text{Reactance of coil, } X_L = 2\pi fL = 2\pi \times 50 \times 318 \times 10^{-3} = 100 \Omega$$

$$\text{Voltage across } L, V_L = IX_L = 2 \times 100 = 200 \text{ V}$$

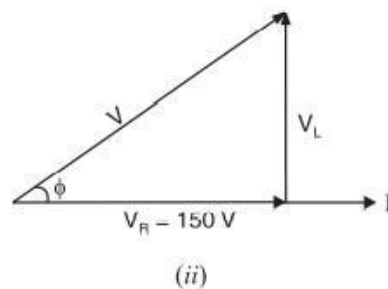
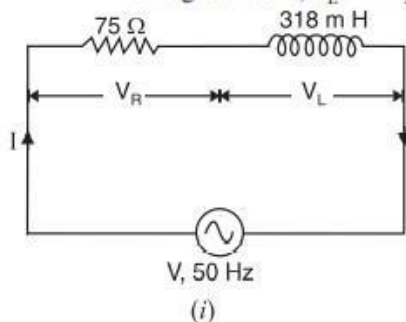


Fig. 12.8

Referring to the phasor diagram of the circuit in Fig. 12.8 (ii),

$$\text{Supply voltage, } V = \sqrt{V_R^2 + V_L^2} = \sqrt{150^2 + 200^2} = \mathbf{250 \text{ V}}$$

Alternatively, $Z = \sqrt{R^2 + X_L^2} = \sqrt{75^2 + 100^2} = 125 \Omega$
 $\therefore V = IZ = 2 \times 125 = 250 \text{ V}$
 Now, $\tan \phi = X_L/R = 100/75 = 1.33$
 \therefore Phase angle, $\phi = \tan^{-1} 1.33 = 53.06^\circ \text{ lag}$

R-C Series A.C. Circuit

Fig. 12.18 shows a resistance of R ohms connected in series with a capacitor of C farad.

- Let $V =$ r.m.s. value of applied voltage
 $I =$ r.m.s. value of the circuit current
 $V_R = IR$ where V_R is in phase with I
 $V_C = IX_C$ where V_C lags I by 90°

Taking current as the reference phasor, the phasor diagram of the circuit can be drawn as shown in Fig. 12.19. The voltage drop $V_R (= IR)$ is in phase with current and is represented in magnitude and direction by the phasor OA . The voltage drop $V_C (= IX_C)$ lags behind the current by 90° and is represented in magnitude and direction by the phasor AB . The applied voltage V is the phasor sum of these two drops *i.e.*

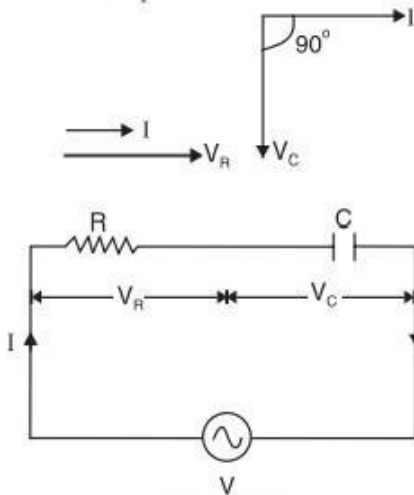


Fig. 12.18

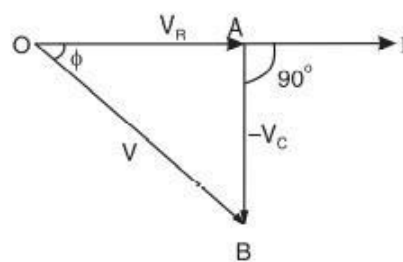


Fig. 12.19

$$V = \sqrt{V_R^2 + (-V_C)^2} = \sqrt{(IR)^2 + (-IX_C)^2} = I \sqrt{R^2 + X_C^2}$$

$$\therefore I = \frac{V}{\sqrt{R^2 + X_C^2}}$$

The quantity $\sqrt{R^2 + X_C^2}$ offers opposition to current flow and is called **impedance** of the circuit.

$$\therefore I = V/Z \text{ where } Z = \sqrt{R^2 + X_C^2}$$

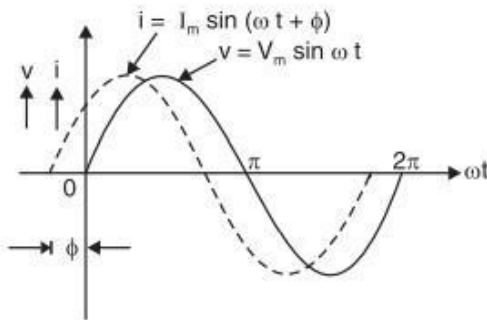
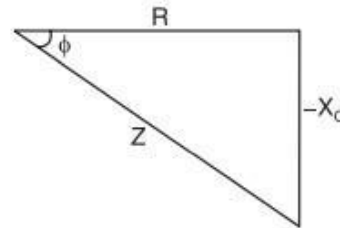


Fig. 12.20



Impedance triangle

Fig. 12.21

(i) **Phase angle.** It is clear from the phasor diagram that circuit current I leads the applied voltage V by ϕ° . This fact is also illustrated in the wave diagram (See Fig. 12.20) and impedance triangle (See Fig. 12.21) of the circuit. The value of the phase angle can be determined as under :

$$\tan \phi = \frac{V_C}{V_R} = \frac{IX_C}{IR} = \frac{X_C}{R}$$

Since current is taken as the reference phasor, negative phase angle implies that voltage lags behind the current. This is the same thing as current leads the voltage.

If the applied voltage is $v = V_m \sin \omega t$, then equation for the circuit current will be :

$$i = I_m \sin(\omega t + \phi) \text{ where } I_m = V_m/Z$$

(ii) **Power.** The equations for voltage and current are :

$$v = V_m \sin \omega t ; i = I_m \sin(\omega t + \phi)$$

\therefore Average power, $P =$ Average of vi

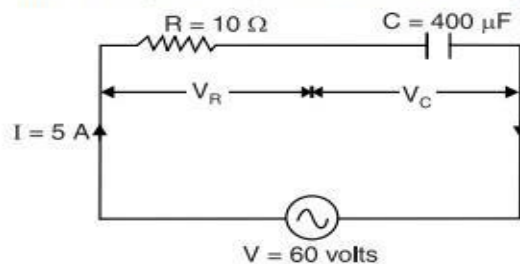
$$= VI \cos \phi$$

Alternatively,

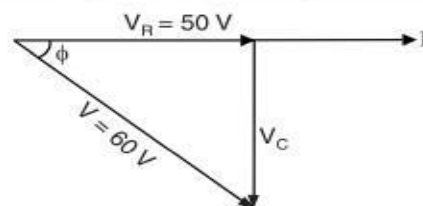
$$P = \text{Power in } R + \text{Power in } C \\ = I^2 R + 0 = IR \times I = IR \times \frac{V}{Z} = VI \times \frac{R}{Z} = VI \cos \phi$$

Example A 10Ω resistor and $400 \mu F$ capacitor are connected in series to a 60-V sinusoidal supply. The circuit current is 5 A . Calculate the supply frequency and phase angle between the current and voltage.

Solution. Fig. 12.22 (i) shows the circuit diagram whereas Fig. 12.22 (ii) shows phasor diagram.



(i)



(ii)

Fig. 12.22

Voltage across R , $V_R = IR = 5 \times 10 = 50 \text{ V}$

Voltage across C , $V_C = \sqrt{V^2 - V_R^2} = \sqrt{60^2 - 50^2} = 33.17 \text{ V}$

Reactance of capacitor, $X_C = V_C / I = 33.17 / 5 = 6.634 \Omega$

$$\therefore \text{Supply frequency, } f = \frac{1}{2\pi C X_C} = \frac{10^6}{2\pi \times 400 \times 6.634} = 60 \text{ Hz}$$

$$\tan \phi = V_C / V_R = 33.17 / 50 = 0.6634$$

$$\therefore \text{Phase angle, } \phi = \tan^{-1} 0.6634 = 33.6^\circ \text{ lead}$$

Example A capacitor of $8 \mu\text{F}$ takes a current of 1A when alternating voltage applied across it is 250V . Calculate (i) frequency of the applied voltage (ii) the resistance to be connected in series with the capacitor to reduce the current in the circuit to 0.5A at the same frequency (iii) phase angle of the resulting circuit.

Solution. (i) Capacitive reactance, $X_C = V/I = 250/1 = 250 \Omega$

$$\therefore \text{Frequency of applied voltage, } f = \frac{1}{2\pi C X_C} = \frac{1}{2\pi \times 8 \times 10^{-6} \times 250} = 79.5 \text{ Hz}$$

(ii) When a resistance is connected in series, the circuit becomes as shown in Fig. 12.23. Since frequency remains the same, the value of X_C is unchanged.

Circuit impedance, $Z = V/I = 250/0.5 = 500 \Omega$

$$\text{Now, } Z^2 = R^2 + X_C^2$$

$$\therefore R = \sqrt{Z^2 - X_C^2} = \sqrt{(500)^2 - (250)^2} = 433 \Omega$$

(iii) Circuit p.f., $\cos \phi = R/Z = 433/500 = 0.866$

\therefore Circuit phase angle, $\phi = \cos^{-1} 0.866 = 30^\circ \text{ lead}$

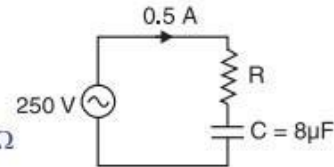


Fig. 12.23

Example A two-element series circuit consumes 700W and has a p.f. of 0.707 leading. If the applied voltage is $v = 141.1 \sin(314t + 30^\circ)$, find the circuit constants.

Solution. Since the circuit p.f. is leading, one circuit element must be a capacitor. Further, power consumed in the circuit is 700W . This suggests that other circuit element is a resistor. Therefore, it is RC series circuit.

R.M.S. value of applied voltage, $V = V_m/\sqrt{2} = 141.1/\sqrt{2} = 100 \text{ volts}$

$$\text{Power consumed, } P = VI \cos \phi \text{ or } 700 = 100 \times I \times 0.707 \therefore I = 10\text{A}$$

$$\text{Also, } P = I^2 R \therefore R = P/I^2 = 700/(10)^2 = 7 \Omega$$

$$\text{Now, } Z = V/I = 100/10 = 10\Omega \therefore X_C = \sqrt{Z^2 - R^2} = \sqrt{(10)^2 - (7)^2} = 7\Omega$$

$$\therefore \text{Capacitance, } C = \frac{1}{2\pi f X_C} = \frac{1}{2\pi \times 50 \times 7} = 450 \times 10^{-6} \text{ F} = 450 \mu\text{F}$$

Example A resistor R in series with a capacitor C is connected to 50Hz , 240V source. Find the value of C so that R absorbs 300W and voltage across R is 100V . Also find the maximum charge and the maximum stored energy in C .

Solution. Supply voltage, $V = 240 \text{ volts}$; $V_R = 100 \text{ volts}$; Power in $R = 300\text{W}$

$$\text{Now, } V^2 = V_R^2 + V_C^2 \text{ or } (240)^2 = (100)^2 + V_C^2 \therefore V_C = 218.17 \text{ volts}$$

$$\text{Value of } R = \frac{V_R^2}{300} = \frac{(100)^2}{300} = 33.33 \Omega$$

$$\text{Circuit current, } I = \frac{V_R}{R} = \frac{100}{33.33} = 3\text{A}$$

$$\text{Now, } X_C = \frac{V_C}{I} = \frac{218.17}{3} = 72.72 \Omega$$

$$\text{or } \frac{1}{2\pi f C} = 72.72 \therefore C = \frac{1}{2\pi \times 50 \times 72.72} = 43.77 \times 10^{-6} \text{ F}$$

$$\text{Max. charge on } C = CV_{C(max)} = (43.77 \times 10^{-6}) \times (\sqrt{2} \times 218.17) = 0.0135 \text{ C}$$

$$\text{Max. energy stored in } C = \frac{1}{2} CV_C^2 = \frac{1}{2} \times 43.77 \times 10^{-6} \times (\sqrt{2} \times 218.17)^2 = 2.08 \text{ J}$$

R-L-C Series A.C. Circuit

This is a general series a.c. circuit. Fig. 12.32 shows R , L and C connected in series across a supply voltage V (r.m.s.). The resulting circuit current is I (r.m.s.).

- ∴ Voltage across R , $V_R = IR$... V_R is in phase with I
- Voltage across L , $V_L = IX_L$... where V_L leads I by 90°
- Voltage across C , $V_C = IX_C$... where V_C lags I by 90°

As before, the phasor diagram is drawn taking current as the reference phasor. In the phasor diagram (See Fig. 12.33), OA represents V_R , AB represents V_L and AC represents V_C . It may be seen that V_L is in phase opposition to V_C . It follows that the circuit can either be effectively inductive or capacitive depending upon which voltage drop (V_L or V_C) is predominant. For the case considered, $V_L > V_C$ so that net voltage drop across L - C combination is $V_L - V_C$ and is represented by AD . Therefore, the applied voltage V is the phasor sum of V_R and $V_L - V_C$ and is represented by OD .

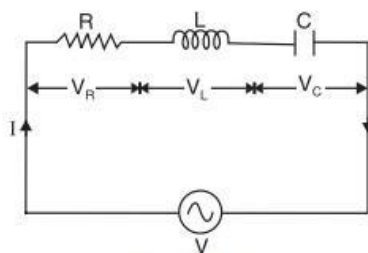


Fig. 12.32

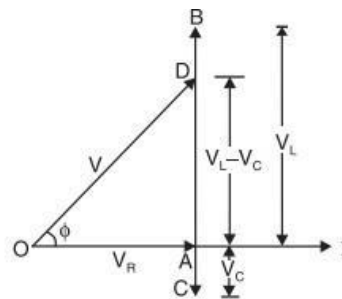


Fig. 12.33

$$\begin{aligned} \therefore V &= \sqrt{V_R^2 + (V_L - V_C)^2} = \sqrt{(IR)^2 + (IX_L - IX_C)^2} \\ &= I \sqrt{R^2 + (X_L - X_C)^2} \\ \therefore I &= \frac{V}{\sqrt{R^2 + (X_L - X_C)^2}} \end{aligned}$$

The quantity $\sqrt{R^2 + (X_L - X_C)^2}$ offers opposition to current flow and is called **impedance** of the circuit.

$$\text{Circuit power factor, } \cos \phi = \frac{R}{Z} = \frac{R}{\sqrt{R^2 + (X_L - X_C)^2}} \quad \dots(i)$$

$$\text{Also, } \tan \phi = \frac{V_L - V_C}{V_R} = \frac{X_L - X_C}{R} \quad \dots(ii)$$

Since X_L , X_C and R are known, phase angle ϕ of the circuit can be determined.

$$\text{Power consumed, } P = VI \cos \phi = I^2 R$$

Three cases of R-L-C series circuit. We have seen that the impedance of a R-L-C series circuit is given by ;

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

- (i) When $X_L - X_C$ is positive (i.e. $X_L > X_C$), phase angle ϕ is positive and the circuit will be inductive. In other words, in such a case, the circuit current I will lag behind the applied voltage V by ϕ ; the value of ϕ being given by eq. (ii) above.
- (ii) When $X_L - X_C$ is negative (i.e. $X_C > X_L$), phase angle ϕ is negative and the circuit is capacitive. That is to say the circuit current I leads the applied voltage V by ϕ ; the value of ϕ being given by eq. (ii) above.
- (iii) When $X_L - X_C$ is zero (i.e. $X_L = X_C$), the circuit is purely resistive. In other words, circuit current I and applied voltage V will be in phase i.e. $\phi = 0^\circ$. The circuit will then have unity power factor.

Example A 230 V, 50 Hz a.c. supply is applied to a coil of 0.06 H inductance and 2.5Ω resistance connected in series with a 6.8 μF capacitor. Calculate (i) impedance (ii) current (iii) phase angle between current and voltage (iv) power factor and (v) power consumed.

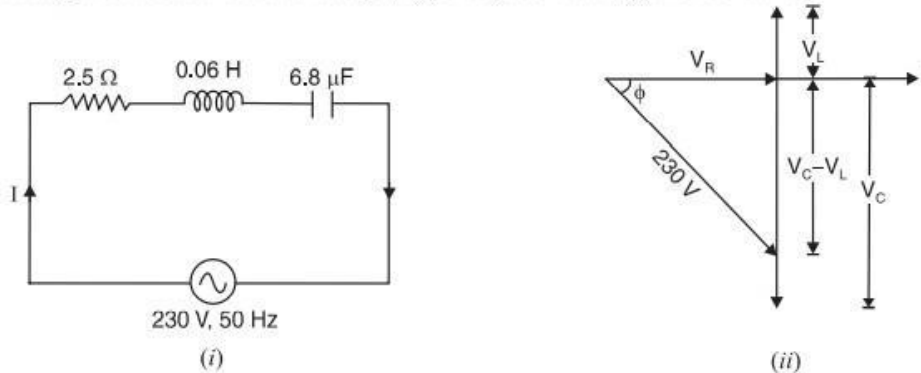


Fig. 12.35

Solution. Fig. 12.35 (i) shows the conditions of the problem.

$$X_L = 2\pi fL = 2\pi \times 50 \times 0.06 = 18.85 \Omega$$

$$X_C = \frac{1}{2\pi f C} = \frac{10^6}{2\pi \times 50 \times 6.8} = 468 \Omega$$

(i) Circuit impedance, $Z = \sqrt{R^2 + (X_L - X_C)^2} = \sqrt{(2.5)^2 + (18.85 - 468)^2} = 449.2 \Omega$

(ii) Circuit current, $I = V/Z = 230/449.2 = 0.512 \text{ A}$

(iii) $\tan \phi = \frac{X_L - X_C}{R} = \frac{18.85 - 468}{2.5} = -179.66$

\therefore Phase angle, $\phi = \tan^{-1} -179.66 = -89.7^\circ = 89.7^\circ \text{ lead}$

The negative sign with ϕ shows that current is leading the voltage [See the phasor diagram in Fig. 12.35 (ii)].

(iv) Power factor, $\cos \phi = \frac{R}{Z} = \frac{2.5}{449.2} = 0.00557 \text{ lead}$

(v) Power consumed, $P = VI \cos \phi = 230 \times 0.512 \times 0.00557 = 0.656 \text{ W}$

Example A coil of p.f. 0.8 is connected in series with a 110 μF capacitor. The supply frequency is 50 Hz. The p.d. across the coil is found to be equal to the p.d. across the capacitor. Calculate the resistance and inductance of the coil.

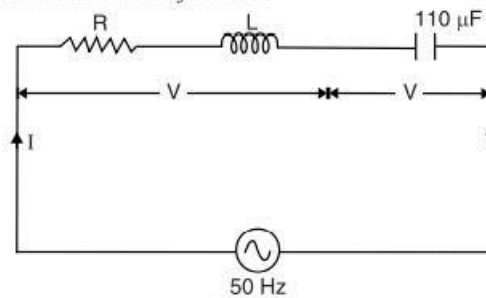


Fig. 12.36

Solution. Fig. 12.36 shows the conditions of the problem.

$$\text{Reactance of capacitor, } X_C = \frac{1}{2\pi f C} = \frac{10^6}{2\pi \times 50 \times 110} = 29 \Omega$$

Now, $I Z_{\text{coil}} = I X_C \therefore Z_{\text{coil}} = X_C = 29 \Omega$

For the coil, $\cos \phi = R/Z_{\text{coil}} \therefore R = Z_{\text{coil}} \cos \phi = 29 \times 0.8 = 23.2 \Omega$

Reactance of coil, $X_L = Z_{\text{coil}} \sin \phi = 29 \times 0.6 = 17.4 \Omega$

\therefore Inductance of coil, $L = \frac{X_L}{2\pi f} = \frac{17.4}{2\pi \times 50} = 0.055 \text{ H}$

***** THE END *****

UNIT – 2 DC AND AC MACHINES

Introduction

The study of the electrical engineering, basically involves the analysis of the energy transfer from one form to another. An electrical machine, deals with the energy transfer either from mechanical to electrical form or from electrical to mechanical form. This process is called **electromechanical energy conversion**.

An electrical machine which converts mechanical energy into an electrical energy is called an **electric generator**. While an electrical machine which converts an electrical energy into the mechanical energy is called an **electrical motor**.

Such electrical machines may be related to an electrical energy of an alternating type called **a.c. machines** or may be related to an electrical energy of direct type called **d.c. machines**.

The d.c. machines are classified as d.c. generators and d.c. motors. The construction of a d.c. machine basically remains same whether it is a generator or a motor.

PRINCIPLE AND OPERATION OF DC GENERATOR

Principle of Operation of a D.C. Generator

All the generators work on a principle of dynamically induced e.m.f. This principle is nothing but the Faraday's law of electromagnetic induction . It states that, 'Whenever the number of magnetic lines of force i.e. flux linking with a conductor or a coil changes, an electromotive force is set up in that conductor or coil.' The change in flux associated with the conductor can exist only when there exists a relative motion between a conductor and the flux. The relative motion can be achieved by rotating conductor with respect to flux or by rotating flux with respect to a conductor. So a voltage gets generated in a conductor, as long as there exists a relative motion between conductor and the flux.

Such an induced e.m.f. which is due to physical movement of coil or conductor with respect to flux or movement of flux with respect to coil or conductor is called **dynamically induced e.m.f.**

Key Point : *So a generating action requires following basic components to exist, i) The conductor or a coil ii) The flux iii) The relative motion between conductor and flux.*

In a practical generator, the conductors are rotated to cut the magnetic flux, keeping flux stationary. To have a large voltage as the output, the number of conductors are connected together in a specific manner, to form a winding. This winding is called **armature winding** of a d.c. machine. The part on which this winding is kept is called **armature** of a d.c. machine. To have the rotation of conductors, the conductors placed on the armature are rotated with the help of some external device. Such an external device is called a **prime mover**. The commonly used prime movers are diesel engines, steam engines, steam turbines, water turbines etc. The necessary magnetic flux is produced by current carrying winding which is called **field winding**. The direction of the induced e.m.f. can be obtained by using Fleming's right hand rule.

Fleming's Right Hand Rule

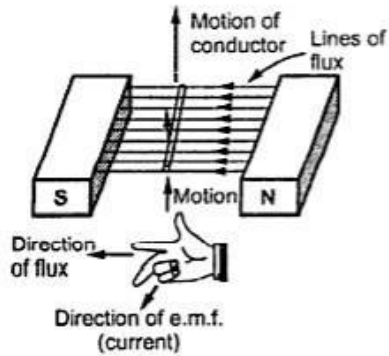


Fig. Fleming's right hand rule

If three fingers of a right hand, namely thumb, index finger and middle finger are outstretched so that everyone of them is at right angles with the remaining two, and if in this position index finger is made to point in the direction of lines of flux, thumb in the direction of the relative motion of the conductor with respect to flux then the outstretched middle finger gives the direction of the e.m.f. induced in the conductor. Visually the rule can be represented as shown in the Fig. 2.6.

This rule mainly gives direction of current which induced e.m.f. in conductor will set up when closed path is provided to it.

Verify the direction of the current through conductor in the four cases shown in the Fig. 2.7 by using Fleming's right hand rule.

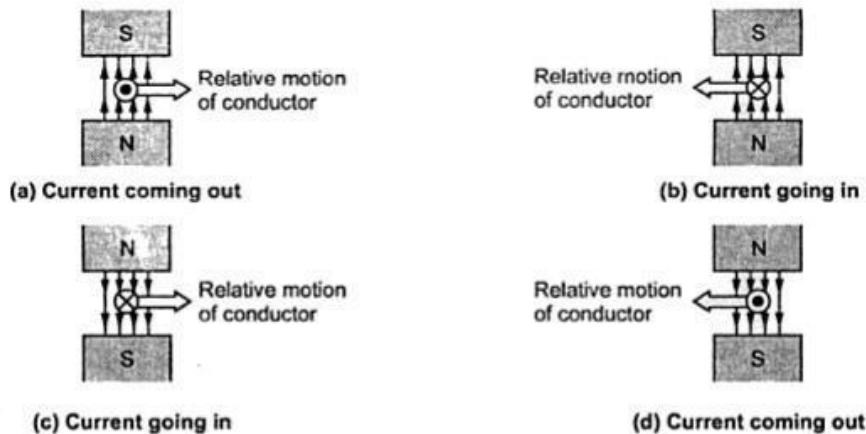


Fig. 2.7 Fleming's right hand rule

Key Point : It can be observed from the Fig. 2.7 that if the direction of relative motion of conductor is reversed keeping flux direction same or if flux direction is reversed keeping direction of relative motion of conductor same then the direction of induced e.m.f. and hence direction of current it sets up in an external circuit gets reversed.

The magnitude of the induced e.m.f. is given by,

$$E = B \times l \times v$$

where

l = Active length of conductor in m.

v = Relative velocity component of conductor in m/s in the direction perpendicular to direction of the flux.

The active length means the length of conductor which is under the influence of magnetic field. In all the cases above, direction of motion of conductor is perpendicular to the plane of the flux.

But if it is not perpendicular then the component of velocity which is perpendicular to the plane of the flux, is only responsible for inducing e.m.f. in the conductor. This is shown in the Fig. 2.8 (a). In this Fig. 2.8 (a), though the velocity is v , its component v' which is perpendicular to the flux lines is only responsible for the induced e.m.f.

If the plane of the rotation of conductor is parallel to the plane of the flux, there will not be any cutting of flux and hence there cannot be any induced e.m.f. in the conductor. This is shown in the Fig. 2.8 (b).

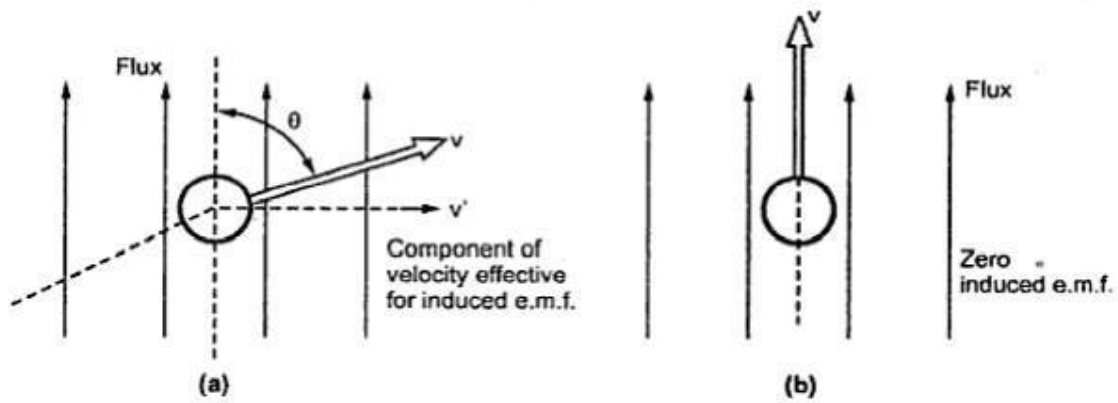


Fig. 2.8

Key Point : So to have an induced e.m.f. in the conductor not only the relative motion between the conductor and the flux is necessary but plane of rotation and plane of flux should not be parallel to each other.

If angle between the plane of rotation and the plane of the flux is θ as measured from the axis of the plane of flux then the induced e.m.f. is given by,

$$E = B l (v \sin\theta) \text{ volts}$$

Where $v \sin\theta$ is the component of velocity which is perpendicular to the plane of flux and hence responsible for the induced e.m.f. This is shown in the Fig. 2.9.

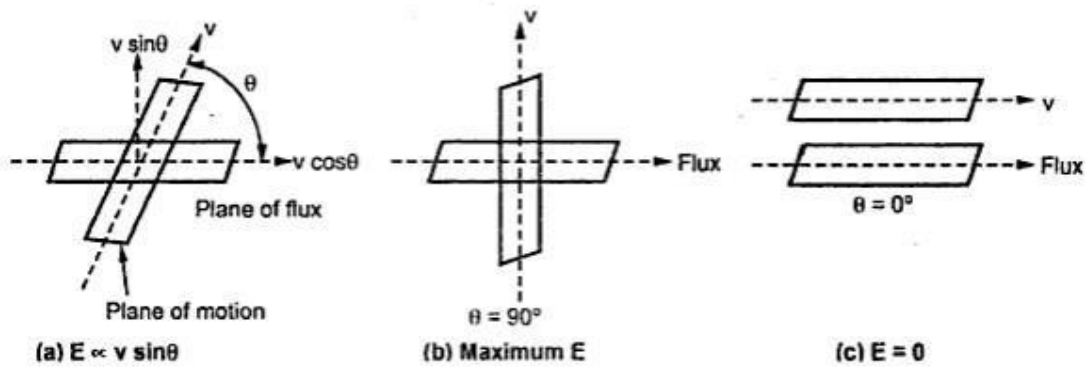


Fig. 2.9

From the equation of the induced e.m.f., it can be seen that the basic nature of the induced e.m.f. in a d.c. generator is purely sinusoidal i.e. alternating. To have d.c. voltage, a device is used in a d.c. generator to convert the alternating e.m.f. to unidirectional e.m.f. This device is called commutator. An alternator is a machine which produces an alternating e.m.f. is without a commutator. So an alternator with a commutator is the basic d.c. generator. Practically there is a difference between the construction of an alternator and a d.c. generator though the basic principle of working is same.

Construction of a Practical D.C. Machine

As stated earlier, whether a machine is d.c. generator or a motor the construction basically remains the same as shown in the Fig. 2.10.

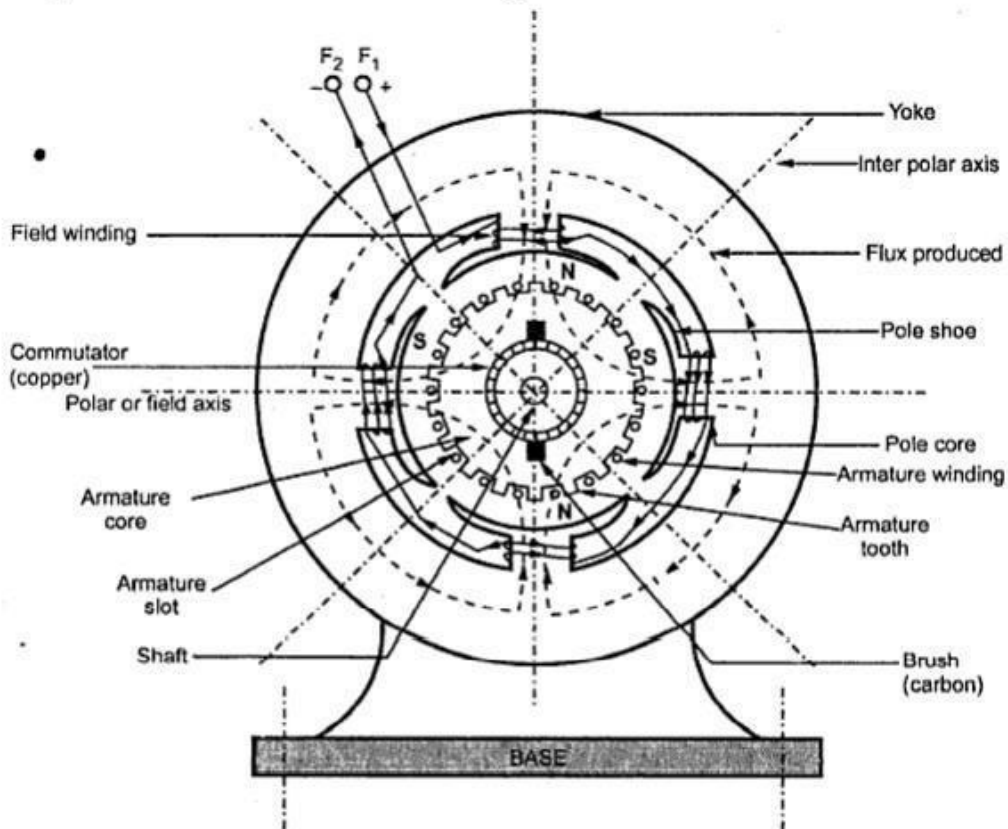


Fig. 2.10 A cross-section of typical d.c. machine

It consists of the following parts :

Yoke

a) Functions :

1. It serves the purpose of outermost cover of the d.c. machine. So that the insulating materials get protected from harmful atmospheric elements like moisture, dust and various gases like SO_2 , acidic fumes etc.
2. It provides mechanical support to the poles.
3. It forms a part of the magnetic circuit. It provides a path of low reluctance for magnetic flux. The low reluctance path is important to avoid wastage of power to provide same flux. Large current and hence the power is necessary if the path has high reluctance, to produce the same flux.

BEEE (Part-A)

b) Choice of material : To provide low reluctance path, it must be made up of some magnetic material. It is prepared by using cast iron because it is cheapest. For large machines rolled steel, cast steel, silicon steel is used which provides high permeability i.e. low reluctance and gives good mechanical strength.

Poles

Each pole is divided into two parts namely, I) Pole core and II) Pole shoe.

This is shown in the Fig. 2.11.

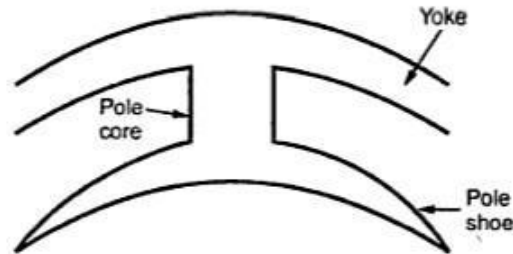


Fig. 2.11 Pole structure

a) Functions of pole core and pole shoe :

1. Pole core basically carries a field winding which is necessary to produce the flux.
2. It directs the flux produced through air gap to armature core, to the next pole.
3. Pole shoe enlarges the area of armature core to come across the flux, which is necessary to produce larger induced e.m.f. To achieve this, pole shoe has been given a particular shape.

b) Choice of material : It is made up of magnetic material like cast iron or cast steel.

As it requires a definite shape and size, laminated construction is used. The laminations of required size and shape are stamped together to get a pole which is then bolted to the yoke.

Field Winding (F1 - F2)

The field winding is wound on the pole core with a definite direction.

a) Functions : To carry current due to which pole core, on which the field winding is placed behaves as an electromagnet, producing necessary flux.

As it helps in producing the magnetic field i.e. exciting the pole as an electromagnet it is called Field winding or Exciting winding.

b) Choice of material : It has to carry current hence obviously made up of some conducting material. So aluminium or copper is the choice. But field coils are required to take any type of shape and bend about pole core and copper has good pliability i.e. it can bend easily. So copper is the proper choice.

Key Point : Field winding is divided into various coils called field coils. These are connected in series with each other and wound in such a direction around pole cores, such that alternate 'N' and 'S' poles are formed.

By using right hand thumb rule for current carrying circular conductor, it can be easily determined that how a particular core is going to behave as 'N' or 'S' for a particular winding direction around it.

Armature

It is further divided into two parts namely,

I) Armature core and II) Armature winding

I) Armature core : Armature core is cylindrical in shape mounted on the shaft. It consists of slots on its periphery and the air ducts to permit the air flow through armature which serves cooling purpose.

a) Functions :

1. Armature core provides house for armature winding i.e. armature conductors.
2. To provide a path of low reluctance to the magnetic flux produced by the field winding.

b) Choice of material : As it has to provide a low reluctance path to the flux, it is made up of magnetic material like cast iron or cast steel.

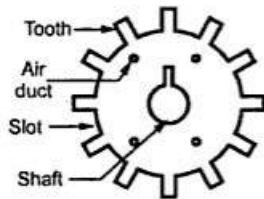


Fig. 2.12 Single circular lamination of armature core

It is made up of laminated construction to keep eddy current loss as low as possible. A single circular lamination used for the construction of the armature core is shown in the Fig. 2.12.

II) Armature winding : Armature winding is nothing but the interconnection of the armature conductors, placed in the slots provided on the armature core periphery. When the armature is rotated, in case of generator, magnetic flux gets cut by

armature conductors and e.m.f. gets induced in them.

a) Functions :

1. Generation of e.m.f. takes place in the armature winding in case of generators.
2. To carry the current supplied in case of d.c. motors.
3. To do the useful work in the external circuit.

b) Choice of material : As armature winding carries entire current which depends on external load, it has to be made up of conducting material, which is copper.

Armature winding is generally former wound. The conductors are placed in the armature slots which are lined with tough insulating material.

2.6.5 Commutator

We have seen earlier that the basic nature of e.m.f. induced in the armature conductors is alternating. This needs rectification in case of d.c. generator, which is possible by a device called commutator.

a) Functions :

1. To facilitate the collection of current from the armature conductors.
2. To convert internally developed alternating e.m.f. to unidirectional (d.c.) e.m.f.
3. To produce unidirectional torque in case of motors.

b) Choice of material : As it collects current from armature, it is also made up of copper segments.

It is cylindrical in shape and is made up of wedge shaped segments of hard drawn, high conductivity copper. These segments are insulated from each other by thin layer of mica. Each commutator segment is connected to the armature conductor by means of copper lug or strip. This connection is shown in the Fig. 2.13.

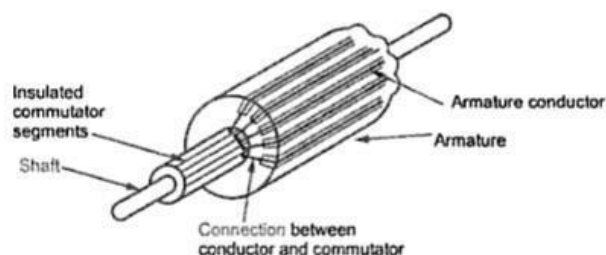


Fig. 2.13 Commutator

Brushes and Brush Gear

Brushes are stationary and resting on the surface of the commutator.

a) **Function** : To collect current from commutator and make it available to the stationary external circuit.

b) **Choice of material** : Brushes are normally made up of soft material like carbon.

Brushes are rectangular in shape. They are housed in brush holders, which are usually of box type. The brushes are made to press on the commutator surface by means of a spring, whose tension can be adjusted with the help of lever. A flexible copper conductor called **pig tail** is used to connect the brush to the external circuit. To avoid wear and tear of commutator, the brushes are made up of soft material like carbon.

Bearings

Ball-bearings are usually used as they are more reliable. For heavy duty machines, roller bearings are preferred.

Types of Armature Winding

We have seen that there are number of armature conductors, which are connected in specific manner as per the requirement, which is called **armature winding**. According to the way of connecting the conductors, armature winding has basically two types namely,

- a) Lap winding
- b) Wave winding

Lap Winding

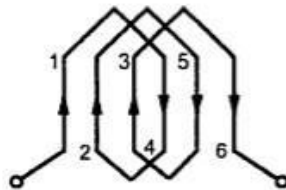


Fig. 2.14 Lap winding

In this case, if connection is started from conductor in slot 1 then connections overlap each other as winding proceeds, till starting point is reached again.

Developed view of part of the armature winding in lap fashion is shown in the Fig. 2.14.

As seen from the Fig. 2.14, there is overlapping of coils while proceeding.

Key Point : Due to such connection, the total number of conductors get divided into 'P' number of parallel paths, where P = number of poles in the machine.

Large number of parallel paths indicate high current capacity of machine hence lap winding is preferred for high current rating generators.

Wave Winding

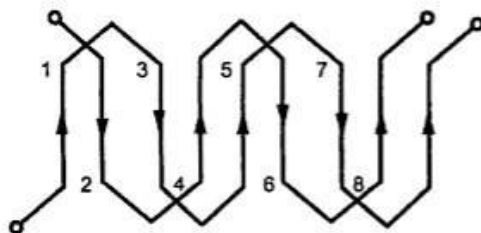


Fig. 2.15 Wave winding

In this type of connection, winding always travels ahead avoiding overlapping. It travels like a progressive wave hence called wave winding. To get an idea of wave winding a part of armature winding in wave fashion is shown in the Fig. 2.15.

Both coils starting from slot 1 and slot 2 are progressing in wave fashion.

Key Point : Due to this type of connection, the total number of conductors get divided into two number of parallel paths always, irrespective of number of poles of the machine. As number of parallel paths are less, it is preferable for low current, high voltage capacity generators.

BEEE (Part-A)

The number of parallel paths in which armature conductors are divided due to lap or wave fashion of connection is denoted as A. So $A = P$ for lap connection and $A = 2$ for wave connection.

Comparison of Lap and Wave Type Winding

Sr. No	Lap winding	Wave winding
1.	Number of parallel paths (A) = poles (P)	Number of parallel paths (A) = 2 (always)
2.	Number of brush sets required is equal to number of poles.	Number of brush sets required is always equal to two.
3.	Preferable for high current, low voltage capacity generators.	Preferable for high voltage, low current capacity generators.
4.	Normally used for generators of capacity more than 500 A.	Preferred for generators of capacity less than 500 A.

E.M.F EQUATION:

E.M.F. Equation of D.C. Generator

Let P = Number of poles of the generator
 ϕ = Flux produced by each pole in webers (Wb)
 N = Speed of armature in r.p.m.
 Z = Total number of armature conductors
 A = Number of parallel paths in which the 'Z' number of conductors are divided

So $A = P$ for lap type of winding
 $A = 2$ for wave type of winding

Now e.m.f. gets induced in the conductor according to Faraday's law of electromagnetic induction. Hence average value of e.m.f. induced in each armature conductor is,

$$e = \text{Rate of cutting the flux} = \frac{d\phi}{dt}$$

Now consider one revolution of conductor. In one revolution, conductor will cut total flux produced by all the poles i.e. $\phi \times P$. While time required to complete one revolution is $\frac{60}{N}$ seconds as speed is N r.p.m.

$$\therefore e = \frac{\phi P}{\frac{60}{N}} = \phi P \frac{N}{60}$$

This is the e.m.f. induced in one conductor. Now the conductors in one parallel path are always in series. There are total Z conductors with A parallel paths, hence $\frac{Z}{A}$ number of conductors are always in series and e.m.f. remains same across all the parallel paths.

\therefore Total e.m.f. can be expressed as,

$$E = \phi P \frac{N}{60} \times \frac{Z}{A} \text{ volts}$$

This is nothing but the e.m.f. equation of a d.c. generator.

So

$$E = \frac{\phi P N Z}{60 A} \text{ e.m.f. equation}$$

$$E = \frac{\phi N Z}{60} \text{ for lap type as } A = P$$

$$E = \frac{\phi P N Z}{120} \text{ for wave type as } A = 2$$

BEEE (Part-A)

➔ **Example** A 4 pole, lap wound, d.c. generator has a useful flux of 0.07 Wb per pole. Calculate the generated e.m.f. when it is rotated at a speed of 900 r.p.m. with the help of prime mover. Armature consists of 440 number of conductors. Also calculate the generated e.m.f. if lap wound armature is replaced by wave wound armature.

Solution : $P = 4$ $Z = 440$ $\phi = 0.07$ Wb and $N = 900$ r.p.m.

$$E = \frac{\phi PNZ}{60A}$$

i) For lap wound,

$$A = P = 4$$

∴

$$E = \frac{\phi NZ}{60} = \frac{0.07 \times 900 \times 440}{60} = 462 \text{ V}$$

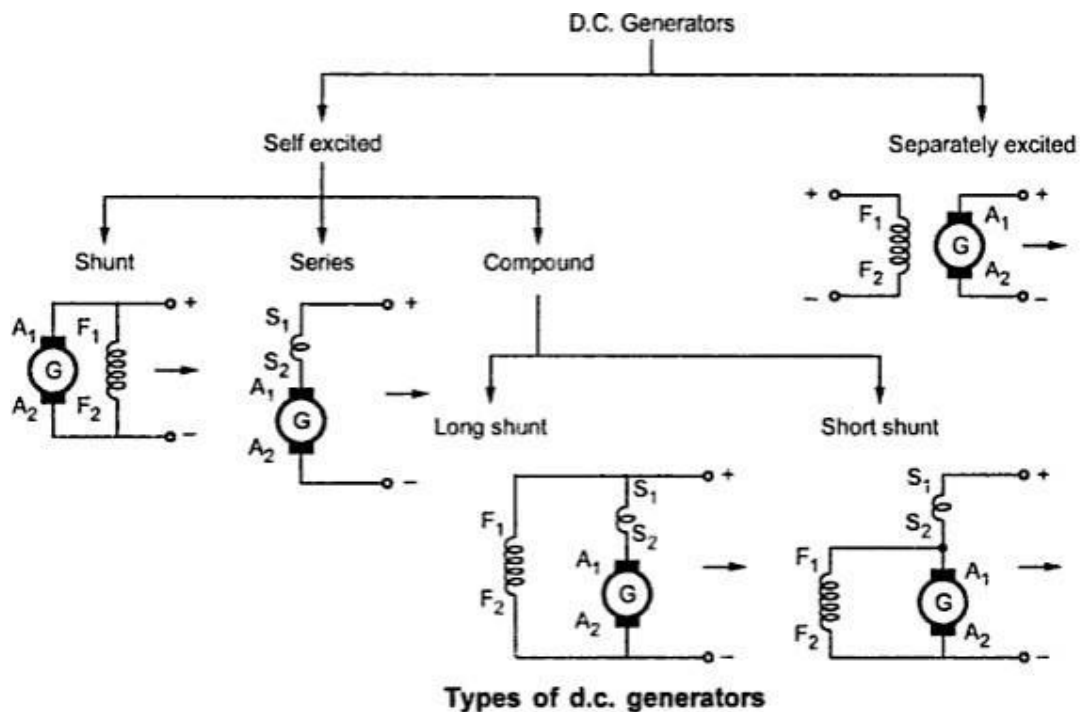
ii) For wave wound

$$A = 2$$

∴

$$E = \frac{\phi PNZ}{120} = \frac{0.07 \times 900 \times 4 \times 440}{120} = 924 \text{ V}$$

TYPES OF DC GENERATORS



OCC CHARACTERISTICS OF DC GENERATOR

No-load saturation Characteristic (E_0/I_f)

It is also known as Magnetic Characteristic or Open-circuit Characteristic (O.C.C.). It shows the relation between the no-load generated e.m.f. in armature, E_0 and the field or exciting current I_f at a given fixed speed. It is just the magnetisation curve for the material of the electromagnets. Its shape is practically the same for all generators whether separately-excited or self-excited.

Separately-excited Generator

No-load Saturation Characteristic ($E_g I_f$)

The arrangement for obtaining the necessary data to plot this curve is shown in Fig. 28.1. The exciting or field current I_f is obtained from an external independent d.c. source. It can be varied from zero upwards by a potentiometer and its value read by an ammeter A connected in the field circuit as shown.

Now, the voltage equation of a d.c. generator is, $E_g = \frac{\Phi ZN}{60} \times \left(\frac{P}{A}\right)$ volt

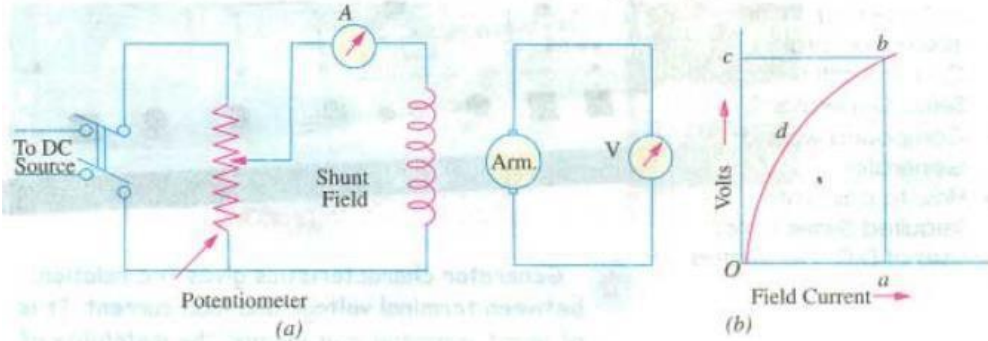


Fig. 28.1

Hence, if speed is constant, the above relation becomes $E = k\Phi$

It is obvious that when I_f is increased from its initial small value, the flux Φ and hence generated e.m.f. E_g increase directly as current so long as the poles are unsaturated. This is represented by the straight portion Od in Fig. 28.1 (b). But as the flux density increases, the poles become saturated, so a greater increase in I_f is required to produce a given increase in voltage than on the lower part of the curve. That is why the upper portion db of the curve Odb bends over as shown.

No-load Curve for Self-excited Generator

The O.C.C. or no-load saturated curves for self-excited generators whether shunt or series-connected, are obtained in a similar way.

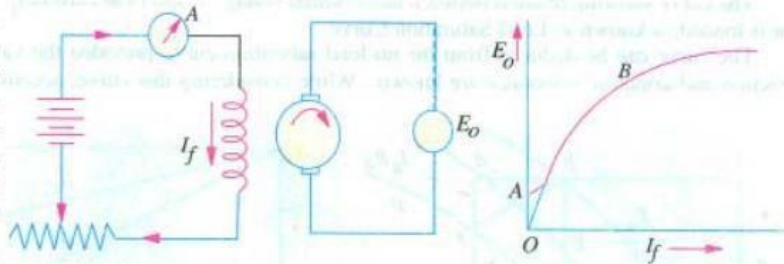
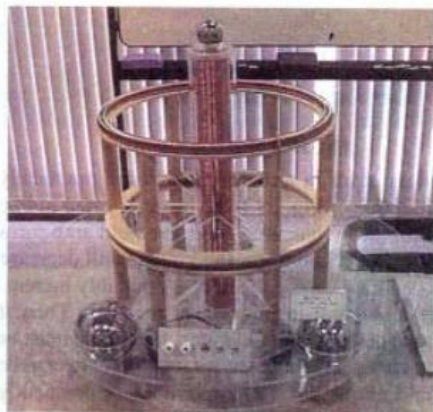


Fig. 28.4

Fig. 28.5

The field winding of the generator (whether *shunt* or *series* wound) is disconnected from the machine and connected to an external source of direct current as shown in Fig. 28.4. The field or exciting current I_f is varied rheostatically and its value read on the ammeter A . The machine is driven at constant speed by the prime mover and the generated e.m.f. on on-load is measured by the voltmeter connected across the armature. I_f is increased by suitable steps (starting from zero) and the corresponding values of E_o are measured. On plotting the relation between I_f and E_o , a curve of this form shown in Fig. 28.5 is obtained.



Self Excited Generator

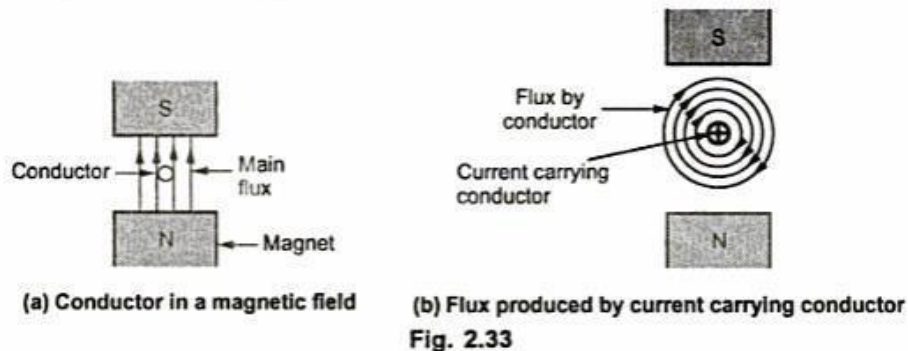
Due to residual magnetism in the poles, some e.m.f. ($= OA$) is generated even when $I_f = 0$. Hence, the curve starts a little way up. The slight curvature at the lower end is due to magnetic inertia. It is seen that the first part of the curve is practically straight. This is due to the fact that at low flux densities, reluctance of iron path being negligible (due to high permeability), total reluctance is given by the air-gap reluctance which is constant. Hence, the flux and consequently, the generated e.m.f. is directly proportional to the exciting current. However, at high flux densities, - where μ is small, iron path reluctance becomes appreciable and straight relation between E and I_f no longer holds good. In other words, after point B , saturation of poles starts. However, the initial slope of the curve is determined by air-gap width.

It should be noted that O.C.C. for a higher speed would lie above this curve and for a lower speed, would lie below it.

Principle of Operation of a D.C. Motor

The principle of operation of a d.c. motor can be stated in a single statement as 'when a current carrying conductor is placed in a magnetic field; it experiences a mechanical force'. In a practical d.c. motor, field winding produces a required magnetic field while armature conductors play a role of a current carrying conductors and hence armature conductors experience a force. As conductors are placed in the slots which are on the periphery, the individual force experienced by the conductors acts as a twisting or turning force on the armature which is called a torque. The torque is the product of force and the radius at which this force acts. So overall armature experiences a torque and starts rotating. Let us study this motoring action in detail.

Consider a single conductor placed in a magnetic field as shown in the Fig. 2.33 (a). The magnetic field is produced by a permanent magnet but in a practical d.c. motor it is produced by the field winding when it carries a current.



Now this conductor is excited by a separate supply so that it carries a current in a particular direction. Consider that it carries a current away from an observer as shown in the Fig. 2.33 (b). Any current carrying conductor produces its own magnetic field around it, hence this conductor also produces its own flux, around. The direction of this flux can be determined by right hand thumb rule. For direction of current considered, the direction of flux around a conductor is clockwise. For simplicity of understanding, the main flux produced by the permanent magnet is not shown in the Fig. 2.33 (b).

Now there are two fluxes present,

1. The flux produced by the permanent magnet called main flux.
2. The flux produced by the current carrying conductor.

These are shown in the Fig. 2.34 (a). From this, it is clear that on one side of the conductor, both the fluxes are in the same direction. In this case, on the left of the conductor there is gathering of the flux lines as two fluxes help each other. As against this, on the right of the conductor, the two fluxes are in opposite direction and hence try to cancel each other. Due to this, the density of the flux lines in this area gets weakened. So on the left, there exists high flux density area while on the right of the conductor there exists low flux density area as shown in the Fig. 2.34 (b).

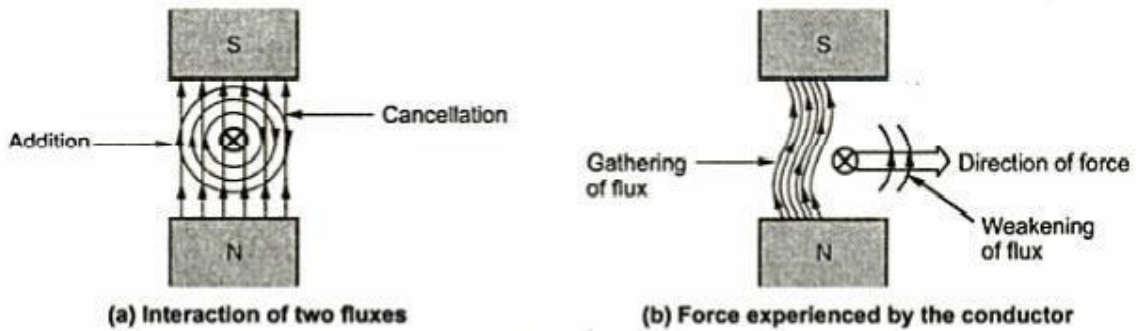


Fig. 2.34

This flux distribution around the conductor acts like a stretched rubber band under tension. This exerts a mechanical force on the conductor which acts from high flux density area towards low flux density area, i.e. from left to right for the case considered as shown in the Fig. 2.34 (b).

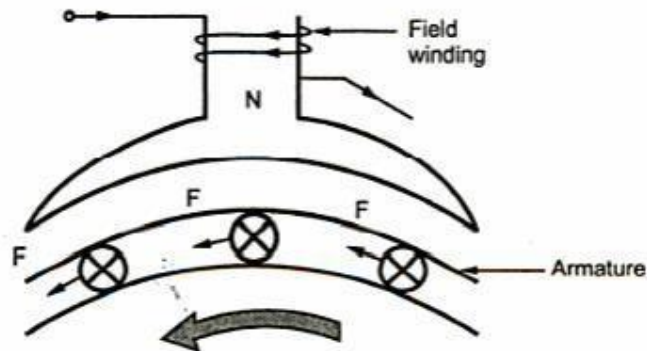


Fig. 2.35 Torque exerted on armature

Key Point : In the practical d.c. motor, the permanent magnet is replaced by a field winding which produces the required flux called main flux and all the armature conductors, mounted on the periphery of the armature drum, get subjected to the mechanical force. Due to this, overall armature experiences a twisting force called torque and armature of the motor starts rotating.

Direction of Rotation of Motor

The magnitude of the force experienced by the conductor in a motor is given by,

$$F = B l I \quad \text{Newtons (N)}$$

B = Flux density due to the flux produced by the field winding.

l = Active length of the conductor.

I = Magnitude of the current passing through the conductor.

The direction of such force i.e. the direction of rotation of a motor can be determined by Fleming's left hand rule. So Fleming's right hand rule is to determine direction of induced e.m.f. i.e. for generating action while Fleming's left hand rule is to determine direction of force experienced i.e. for motoring action.

Fleming's Left Hand Rule

The rule states that, 'Outstretch the three fingers of the left hand namely the first finger, middle finger and thumb such that they are mutually perpendicular to each other. Now point the first finger in the direction of magnetic field and the middle finger in the direction of the current then the thumb gives the direction of the force experienced by the conductor'.

The Fleming's left hand rule can be diagrammatically shown as in the Fig. 2.36.

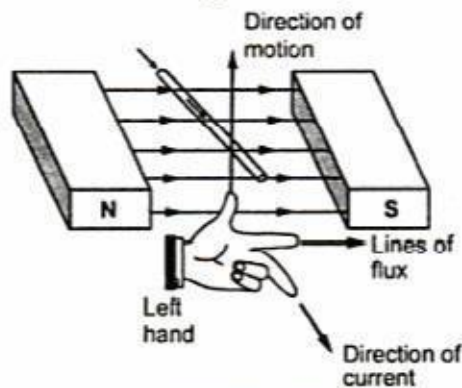


Fig. 2.36 Fleming's left hand rule

Apply the rule to crosscheck the direction of force experienced by a single conductor, placed in the magnetic field, shown in the Fig. 2.37 (a), (b), (c) and (d).

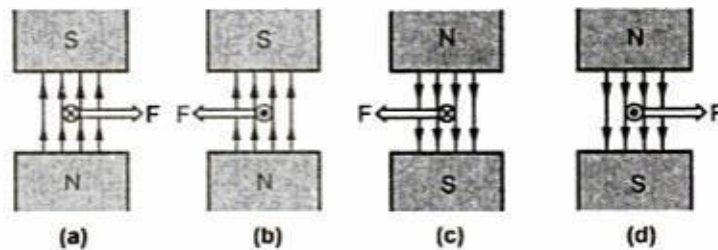


Fig. 2.37 Direction of force experienced by conductor

It can be seen from the Fig. 2.37 that if the direction of the main field in which current carrying conductor is placed, is reversed, force experienced by the conductor reverses its direction. Similarly keeping main flux direction unchanged, the direction of current passing through the conductor is reversed, the force experienced by the conductor reverses its direction. However if both the directions are reversed, the direction of the force experienced remains the same.

Key Point: So in a practical motor, to reverse its direction of rotation, either direction of main field produced by the field winding is reversed or direction of the current passing through the armature is reversed.

The direction of the main field can be reversed by changing the direction of current passing through the field winding, which is possible by interchanging the polarities of supply which is given to the field winding. In short, to have a motoring action two fluxes must exist, the interaction of which produces a torque.

Significance of Back E.M.F.

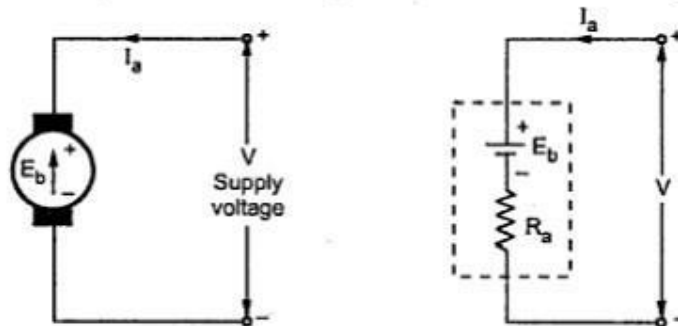
It is seen in the generating action, that when a conductor cuts the lines of flux, e.m.f. gets induced in the conductor. The question is obvious that in a d.c. motor, after a motoring action, armature starts rotating and armature conductors cut the main flux. So is there a generating action existing in a motor? The answer to this question is 'Yes'.

After a motoring action, there exists a generating action. There is an induced e.m.f. in the rotating armature conductors according to Faraday's law of electromagnetic induction. This induced e.m.f. in the armature always acts in the opposite direction of the supply voltage. This is according to the **Lenz's law** which states that the direction of the induced e.m.f. is always so as to oppose the cause producing it. In a d.c. motor, electrical input i.e. the supply voltage is the cause and hence this induced e.m.f. opposes the supply voltage. This e.m.f. tries to set up a current through the armature which is in the opposite direction to that, which supply voltage is forcing through the conductor.

So as this e.m.f. always opposes the supply voltage, it is called **back e.m.f.** and denoted as E_b . Though it is denoted as E_b , basically it gets generated by the generating action which we have seen earlier in case of generators. So its magnitude can be determined by the e.m.f. equation which is derived earlier. So,

$$E_b = \frac{\phi P N Z}{60 A} \text{ volts}$$

where all symbols carry the same meaning as seen earlier in case of generators.



BEEE (Part-A)

Voltage Equation of a D.C. Motor

In case of a generator, generated e.m.f. has to supply armature resistance drop and remaining part is available across the load as a terminal voltage. But in case of d.c. motor, supply voltage V has to overcome back e.m.f. E_b which is opposing V and also various drops as armature resistance drop $I_a R_a$, brush drop etc. Infact the electrical work done in overcoming the back e.m.f. gets converted into the mechanical energy developed in the armature. Hence the voltage equation of a d.c. motor can be written as,

$$V = E_b + I_a R_a + \text{brush drop}$$

Neglecting the brush drop, the generalised voltage equation is,

$$V = E_b + I_a R_a$$

The back e.m.f. is always less than supply voltage ($E_b < V$). But R_a is very small hence under normal running conditions, the difference between back e.m.f. and supply voltage is very small. The net voltage across the armature is the difference between the supply voltage and back e.m.f. which decides the armature current. Hence from the voltage equation we can write,

$$I_a = \frac{V - E_b}{R_a}$$

Key Point : Voltage equation gets changed a little bit depending upon the type of the motor, which is discussed later.

Power Equation of a D.C. Motor

The voltage equation of a d.c. motor is given by,

$$V = E_b + I_a R_a$$

Multiplying both sides of the above equation by I_a we get,

$$VI_a = E_b I_a + I_a^2 R_a$$

This equation is called **power equation** of a d.c. motor.

VI_a = Net electrical power input to the armature measured in watts.

$I_a^2 R_a$ = Power loss due to the resistance of the armature called **armature copper loss**.

So difference between VI_a and $I_a^2 R_a$ i.e. input - losses gives the output of the armature.

So $E_b I_a$ is called **electrical equivalent of gross mechanical power developed by the armature**. This is denoted as P_m .

\therefore Power input to the armature - Armature copper loss = Gross mechanical power developed in the armature.

Condition for Maximum Power

For a motor from power equation it is known that,

$$\begin{aligned} P_m &= \text{Gross mechanical power developed} = E_b I_a \\ &= VI_a - I_a^2 R_a \end{aligned}$$

For maximum P_m , $\frac{dP_m}{dI_a} = 0$

$$\therefore 0 = V - 2I_a R_a$$

$$\therefore I_a = \frac{V}{2R_a} \quad \text{i.e.} \quad I_a R_a = \frac{V}{2}$$

Substituting in voltage equation,

$$V = E_b + I_a R_a = E_b + \frac{V}{2}$$

$\therefore E_b = \frac{V}{2}$ Condition for maximum power

Key Point : This is practically impossible to achieve as for this E_b , current required is much more than its normal rated value. Large heat will be produced and efficiency of motor will be less than 50 %.

Torque Equation of a D.C. Motor

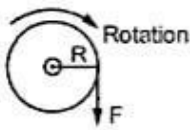


Fig. 2.39

It is seen that the turning or twisting force about an axis is called torque. Consider a wheel of radius R meters acted upon by a circumferential force F newtons as shown in the Fig. 2.39.

The wheel is rotating at a speed of N r.p.m.

Then angular speed of the wheel is,

$$\omega = \frac{2\pi N}{60} \text{ rad/sec}$$

So workdone in one revolution is,

$$\begin{aligned} W &= F \times \text{distance travelled in one revolution} \\ &= F \times 2\pi R \text{ joules} \end{aligned}$$

And

$$\begin{aligned} P &= \text{Power developed} = \frac{\text{Workdone}}{\text{Time}} \\ &= \frac{F \times 2\pi R}{\text{Time for 1 rev}} = \frac{F \times 2\pi R}{\left(\frac{60}{N}\right)} = (F \times R) \times \left(\frac{2\pi N}{60}\right) \end{aligned}$$

$$\therefore P = T \times \omega \text{ watts}$$

Where T = Torque in N - m

ω = Angular speed in rad/sec.

Let T_a be the gross torque developed by the armature of the motor. It is also called **armature torque**. The gross mechanical power developed in the armature is $E_b I_a$, as seen from the power equation. So if speed of the motor is N r.p.m. then,

Power in armature = Armature torque $\times \omega$

$$\therefore E_b I_a = T_a \times \frac{2\pi N}{60}$$

but E_b in a motor is given by,

$$E_b = \frac{\phi P N Z}{60 A}$$

$$\therefore \frac{\phi P N Z}{60 A} \times I_a = T_a \times \frac{2\pi N}{60}$$

$$T_a = \frac{1}{2\pi} \phi I_a \times \frac{PZ}{A}$$

$$T_a = 0.159 \phi I_a \cdot \frac{PZ}{A} \text{ N-m}$$

This is the torque equation of a d.c. motor.

PERFORMANCE CHARACTERISTICS OF DC MOTOR

Characteristics of D.C. Shunt Motor

i) Torque - armature current characteristics

For a d.c. motor $T \propto \phi I_a$

For a constant values of R_{sh} and supply voltage V , I_{sh} is also constant and hence flux is also constant.

\therefore $T_a \propto I_a$

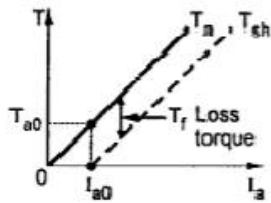


Fig. 2.45 T Vs I_a for shunt motor

The equation represents a straight line, passing through the origin, as shown in the Fig. 2.45. Torque increases linearly with armature current. It is seen earlier that armature current is decided by the load. So as load increases, armature current increases, increasing the torque developed linearly.

Now if shaft torque is plotted against armature current, it is known that shaft torque is less than the armature torque and the difference between the two is loss torque T_f as shown. On no load $T_{sh} = 0$ but armature torque is present which is just enough to overcome stray losses shown as T_{a0} . The current required is I_{a0} on no load to produce T_{a0} and hence T_{sh} graph has an intercept of I_{a0} on the current axis.

To generate high starting torque, this type of motor requires a large value of armature current at start. This may damage the motor hence d.c. shunt motors can develop moderate starting torque and hence suitable for such applications where starting torque requirement is moderate.

ii) Speed - armature current characteristics

From the speed equation we get,

$$N \propto \frac{V - I_a R_a}{\phi}$$

$\propto V - I_a R_a$ as ϕ is constant.

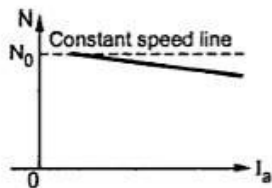


Fig. 2.46 N Vs I_a for shunt motor

So as load increases, the armature current increases and hence drop $I_a R_a$ also increases.

Hence for constant supply voltage, $V - I_a R_a$ decreases and hence speed reduces. But as R_a is very small, for change in I_a from no load to full load, drop $I_a R_a$ is very small and hence drop in speed is also not significant from no load to full load.

So the characteristics is slightly dropping as shown in the Fig. 2.46.

But for all practical purposes these type of motors are considered to be a constant speed motors.

iii) Speed - torque characteristics

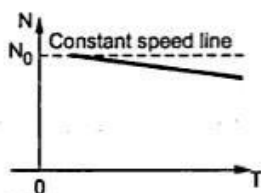


Fig. 2.47 N Vs T for shunt motor

These characteristics can be derived from the above two characteristics. This graph is similar to speed-armature current characteristics as torque is proportional to the armature current. This curve shows that the speed almost remains constant though torque changes from no load to full load conditions. This is shown in the Fig. 2.47.

BEEE (Part-A)
Characteristics of D.C. Series Motor

i) Torque - armature current characteristics

In case of series motor the series field winding is carrying the entire armature current. So flux produced is proportional to the armature current.

$$\therefore \phi \propto I_a$$

Hence

$$T_a \propto \phi I_a \propto I_a^2$$

Thus torque in case of series motor is proportional to the square of the armature current. This relation is parabolic in nature as shown in the Fig. 2.48.

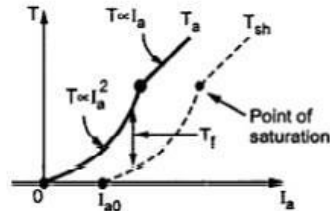


Fig. 2.48 T Vs I_a for series motor

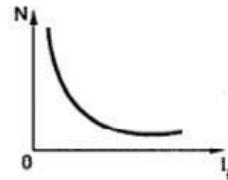


Fig. 2.49 N Vs I_a for series motor

As load increases, armature current increases and torque produced increases proportional to the square of the armature current upto a certain limit.

As the entire I_a passes through the series field, there is a property of an electromagnet called saturation, may occur. Saturation means though the current through the winding increases, the flux produced remains constant. Hence after saturation the characteristics take the shape of straight line as flux becomes constant, as shown. The difference between T_a and T_{sh} is loss torque T_f which is also shown in the Fig. 2.48.

At start as $T \propto I_a^2$, these types of motors can produce high torque for small amount of armature current hence the series motors are suitable for the applications which demand high starting torque.

ii) Speed - armature current characteristics

From the speed equation we get,

$$N \propto \frac{E_b}{\phi} \propto \frac{V - I_a R_a - I_a R_{se}}{I_a} \quad \text{as } \phi \propto I_a \text{ in case of series motor}$$

Now the values of R_a and R_{se} are so small that the effect of change in I_a on speed overrides the effect of change in $V - I_a R_a - I_a R_{se}$ on the speed.

Hence in the speed equation, $E_b \cong V$ and can be assumed constant. So speed equation reduces to,

$$N \propto \frac{1}{I_a}$$

So speed-armature current characteristics is rectangular hyperbola type as shown in the Fig. 2.49.

iii) Speed - torque characteristics

In case of series motors, $T \propto I_a^2$ and $N \propto \frac{1}{I_a}$

Hence we can write,
$$N \propto \frac{1}{\sqrt{T}}$$

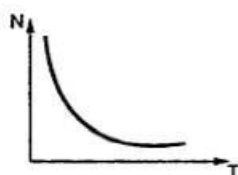


Fig. 2.50 N Vs T for series motor

Thus as torque increases when load increases, the speed decreases. On no load, torque is very less and hence speed increases to dangerously high value. Thus the nature of the speed-torque characteristics is similar to the nature of the speed-armature current characteristics.

The speed-torque characteristics of a series motor is shown in the Fig. 2.50.

Why Series Motor is Never Started on No Load ?

It is seen earlier that motor armature current is decided by the load. On light load or no load, the armature current drawn by the motor is very small.

In case of a d.c. series motor, $\phi \propto I_a$ and

on no load as I_a is small hence flux produced is also very small.

According to speed equation,

$$N \propto \frac{1}{\phi} \quad \text{as } E_b \text{ is almost constant.}$$

So on very light load or no load as flux is very small, the motor tries to run at dangerously high speed which may damage the motor mechanically. This can be seen from the speed-armature current and the speed-torque characteristics that on low armature current and low torque condition motor shows a tendency to rotate with dangerously high speed.

This is the reason why series motor should never be started on light loads or no load conditions. For this reason it is not selected for belt drives as breaking or slipping of belt causes to throw the entire load off on the motor and made to run motor with no load which is dangerous.

Characteristics of D.C. Compound Motor

Compound motor characteristics basically depends on the fact whether the motor is cumulatively compound or differential compound. All the characteristics of the compound motor are the combination of the shunt and series characteristic.

Cumulative compound motor is capable of developing large amount of torque at low speeds just like series motor. However it is not having a disadvantage of series motor even at light or no load. The shunt field winding produces the definite flux and series flux helps the shunt field flux to increase the total flux level.

So cumulative compound motor can run at a reasonable speed and will not run with dangerously high speed like series motor, on light or no load condition.

In differential compound motor, as two fluxes oppose each other, the resultant flux decreases as load increases, thus the machine runs at a higher speed with increase in the load. This property is dangerous as on full load, the motor may try to run with dangerously high speed. So differential compound motor is generally not used in practice.

The various characteristics of both the types of compound motors cumulative and the differential are shown in the Fig. 2.51 (a), (b) and (c).

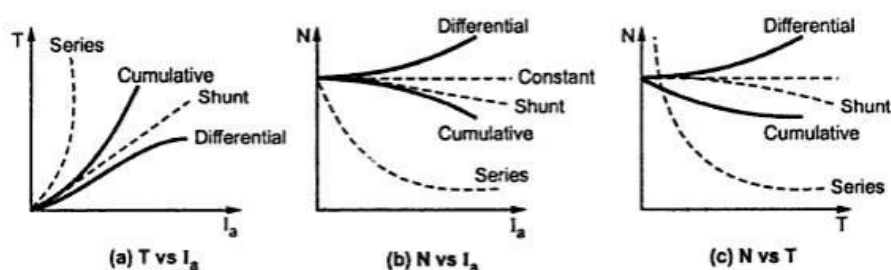


Fig. 2.51 Characteristics of d.c. compound motor

The exact shape of these characteristics depends on the relative contribution of series and shunt field windings. If the shunt field winding is more dominant then the characteristics take the shape of the shunt motor characteristics. While if the series field winding is more dominant then the characteristics take the shape of the series characteristics.

Applications of D.C. Motors

Instead of just stating the applications, the behaviour of the various characteristics like speed, starting torque etc., which makes the motor more suitable for the applications, is also stated in the Table 2.1.

Type of motor	Characteristics	Applications
Shunt	Speed is fairly constant and medium starting torque.	1) Blowers and fans 2) Centrifugal and reciprocating pumps 3) Lathe machines 4) Machine tools 5) Milling machines 6) Drilling machines.
Series	High starting torque. No load condition is dangerous. Variable speed.	1) Cranes 2) Hoists, Elevators 3) Trolleys 4) Conveyors 5) Electric locomotives.
Cumulative compound	High starting torque. No load condition is allowed.	1) Rolling mills 2) Punches 3) Shears 4) Heavy planers 5) Elevators.
Differential compound	Speed increases as load increases.	Not suitable for any practical application.

Table 2.1

Losses in a D.C. Machine

The various losses in a d.c. machine whether it is a motor or a generator are classified into three groups as :

1. Copper losses
2. Iron or core losses
3. Mechanical losses.

Copper Losses

The copper losses are the losses taking place due to the current flowing in a winding. There are basically two windings in a d.c. machine namely armature winding and field winding. The copper losses are proportional to the square of the current flowing through these windings. Thus the various copper losses can be given by,

$$\text{Armature copper loss} = I_a^2 R_a$$

where R_a = Armature winding resistance

and I_a = Armature current

$$\text{Shunt field copper loss} = I_{sh}^2 R_{sh}$$

where R_{sh} = Shunt field winding resistance

and I_{sh} = Shunt field current

$$\text{Series field copper loss} = I_{se}^2 R_{se}$$

where R_{se} = Series field winding resistance

and I_{se} = Series field current

In a compound d.c. machine, both shunt and series field copper losses are present. In addition to the copper losses, there exists brush contact resistance drop. But this drop is usually included in the armature copper loss.

Iron or Core Losses

These losses are also called magnetic losses. These losses include hysteresis loss and eddy current loss.

The hysteresis loss is proportional to the frequency and the maximum flux density B_m in the air gap and is given by,

$$\text{Hysteresis loss} = \eta B_m^{1.6} f V \text{ watts}$$

η = Steinmetz hysteresis coefficient

where V = Volume of core in m^3

f = Frequency of magnetic reversals

This loss is basically due to reversal of magnetisation of the armature core.

The eddy current loss exists due to eddy currents. When armature core rotates, it cuts the magnetic flux and e.m.f. gets induced in the core. This induced e.m.f. sets up eddy currents which cause the power loss. This loss is given by,

$$\text{Eddy current loss} = K B_m^2 f^2 t^2 V \text{ watts}$$

where K = Constant

t = Thickness of each lamination

V = Volume of core

f = Frequency of magnetic reversals

The hysteresis loss is minimised by selecting the core material having low hysteresis coefficient. While eddy current loss is minimised by selecting the laminated construction for the core.

These losses are almost constant for the d.c. machines.

Mechanical Losses

These losses consist of friction and windage losses. Some power is required to overcome mechanical friction and wind resistance at the shaft. This loss is nothing but the friction and windage loss. The mechanical losses are also constant for a d.c. machine.

The magnetic and mechanical losses together are called **stray losses**. For the shunt and compound d.c. machines where field current is constant, field copper losses are also constant. Thus stray losses along with constant field copper losses are called **constant losses**. While the armature current is dependent on the load and thus armature copper losses are called **variable losses**.

Thus for a d.c. machine,

$$\text{Total losses} = \text{Constant losses} + \text{Variable losses}$$

The power flow and energy transformation diagrams at various stages, which takes place in a d.c. machine are represented diagrammatically in Fig. 2.60 (a) and (b).

Efficiency of a D.C. Machine

For a d.c. machine, its overall efficiency is given by,

$$\% \eta = \frac{\text{Total output}}{\text{Total input}} \times 100$$

Let P_{out} = Total output of a machine

P_{in} = Total input of a machine

P_{cu} = Variable losses

P_i = Constant losses

then $P_{\text{in}} = P_{\text{out}} + P_{\text{cu}} + P_i$

$$\therefore \% \eta = \frac{P_{\text{out}}}{P_{\text{in}}} \times 100 = \frac{P_{\text{out}}}{P_{\text{out}} + \text{losses}} \times 100$$

$$\therefore \% \eta = \frac{P_{\text{out}}}{P_{\text{in}} + P_{\text{cu}} + P_i} \times 100$$

PRINCIPLE AND OPERATION OF SINGLE-PHASE TRANSFORMER

Introduction

The main advantage of alternating currents over direct currents is that, the alternating currents can be easily transferable from low voltage to high or high voltage to low. Alternating voltages can be raised or lowered as per requirements in the different stages of electrical network as generation, transmission, distribution and utilization. This is possible with a static device called **transformer**. The transformer works on the principle of mutual induction. It transfers an electric energy from one circuit to other when there is no electrical connection between the two circuits. Thus we can define transformer as below :

Key Point: *The transformer is a static piece of apparatus by means of which an electrical power is transformed from one alternating current circuit to another with the desired change in voltage and current, without any change in the frequency.*

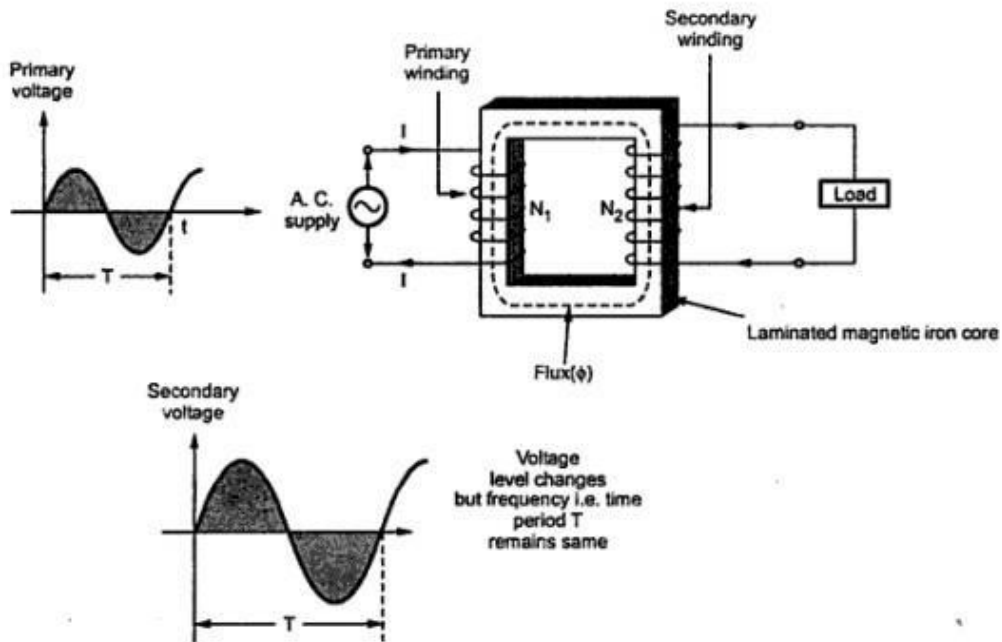
Principle of Working

The principle of mutual induction states that when two coils are inductively coupled and if current in one coil is changed uniformly then an e.m.f. gets induced in the other coil. This e.m.f. can drive a current, when a closed path is provided to it. The transformer works on the same principle. In its elementary form, it consists of two inductive coils

BEEE (Part-A)

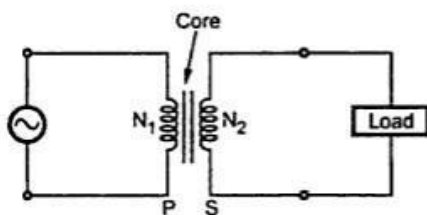
which are electrically separated but linked through a common magnetic circuit. The two coils have high mutual inductance. The basic transformer is shown in the Fig. 3.2.

One of the two coils is connected to a source of alternating voltage. This coil in which electrical energy is fed with the help of source is called **primary winding (P)**. The other winding is connected to load. The electrical energy transformed to this winding is drawn out to the load.



Basic transformer

This winding is called **secondary winding (S)**. The primary winding has N_1 number of turns while the secondary winding has N_2 number of turns. Symbolically the transformer is indicated as shown in the Fig. 3.3.



Symbolic representation

When primary winding is excited by an alternating voltage, it circulates an alternating current. This current produces an alternating flux (ϕ) which completes its path through common magnetic core as shown dotted in the Fig. 3.2. Thus an alternating flux links with the

secondary winding. As the flux is alternating, according to Faraday's law of an electromagnetic induction, mutually induced e.m.f. gets developed in the secondary

winding. If now load is connected to the secondary winding, this e.m.f. drives a current through it.

Thus though there is no electrical contact between the two windings, an electrical energy gets transferred from primary to the secondary.

Key Point: *The frequency of the mutually induced e.m.f. is same as that of the alternating source which is supplying energy to the primary winding.*

Can D.C. Supply be used for Transformers ?

The d.c. supply can not be used for the transformers.

The transformer works on the principle of mutual induction, for which current in one coil must change uniformly. If d.c. supply is given, the current will not change due to constant supply and transformer will not work.

Practically winding resistance is very small. For d.c., the inductive reactance X_L is zero as d.c. has no frequency. So total impedance of winding is very low for d.c. Thus winding will draw very high current if d.c. supply is given to it. This may cause the burning of windings due to extra heat generated and may cause permanent damage to the transformer.

There can be saturation of the core due to which transformer draws very large current from the supply when connected to d.c.

Thus d.c. supply should not be connected to the transformers.

Predetermination of Efficiency and Regulation

The efficiency and regulation of a transformer on any load condition and at any power factor condition can be predetermined by indirect loading method. In this method, the actual load is not used on transformer. But the equivalent circuit parameters of a transformer are determined by conducting two tests on a transformer which are,

1. Open circuit test (O.C. test)
2. Short circuit test (S.C. test)

The parameters calculated from these test results are effective in determining the regulation and efficiency of a transformer at any load and power factor condition, without actually loading the transformer. The advantage of this method is that without much power loss the tests can be performed and results can be obtained. Let us discuss in detail how to perform these tests and how to use the results to calculate equivalent circuit parameters.

Open Circuit Test (O.C. Test)

The experimental circuit to conduct O.C. test is shown in the Fig. 3.25.

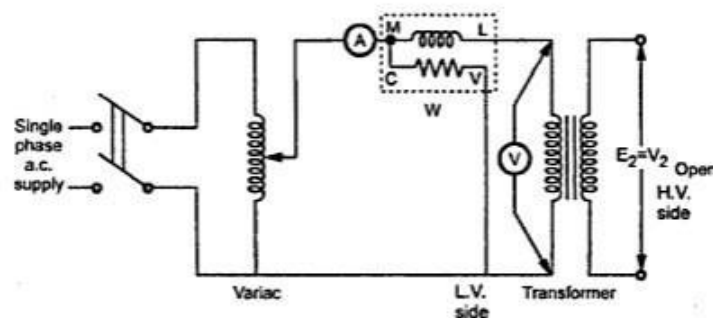


Fig. 3.25 Experimental circuit for O.C. test

The transformer primary is connected to a.c. supply through ammeter, wattmeter and variac. The secondary of transformer is kept open. Usually low voltage side is used as primary and high voltage side as secondary to conduct O.C. test.

BEEE (Part-A)

The primary is excited by rated voltage, which is adjusted precisely with the help of a variac. The wattmeter measures input power. The ammeter measures input current. The voltmeter gives the value of rated primary voltage applied at rated frequency.

Sometimes a voltmeter may be connected across secondary to measure secondary voltage which is $V_2 = E_2$ when primary is supplied with rated voltage. As voltmeter resistance is very high, though voltmeter is connected, secondary is treated to be open circuit as voltmeter current is always negligibly small.

When the primary voltage is adjusted to its rated value with the help of variac, readings of ammeter and wattmeter are to be recorded.

The observation table is as follows.

V_o volts	I_o amperes	W_o watts
Rated		

$$V_o = \text{Rated voltage}$$

$$W_o = \text{Input power}$$

$$I_o = \text{Input current} = \text{No load current}$$

As transformer secondary is open, it is on no load. So current drawn by the primary is no load current I_o . The two components of this no load current are,

$$I_m = I_o \sin \phi_o$$

$$I_c = I_o \cos \phi_o$$

where $\cos \phi_o = \text{No load power factor}$

And hence power input can be written as,

$$W_o = V_o I_o \cos \phi_o$$

The phasor diagram is shown in the Fig. 3.26.

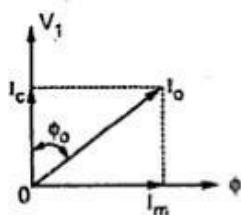


Fig. 3.26

As secondary is open, $I_2 = 0$. Thus its reflected current on primary I_2' is also zero. So we have primary current $I_1 = I_o$. The transformer no load current is always very small, hardly 2 to 4 % of its full load value. As $I_2 = 0$, secondary copper losses are zero. And $I_1 = I_o$ is very low hence copper losses on primary are also very very low. Thus the total copper losses in O.C. test are negligibly small. As against this the input voltage is rated at rated frequency hence flux density in the core is at its maximum value. Hence iron losses are at rated voltage. As output power is

zero and copper losses are very low, the total input power is used to supply iron losses.

This power is measured by the wattmeter i.e. W_o . Hence the wattmeter in O.C. test gives iron losses which remain constant for all the loads.

$$\therefore W_o = P_i = \text{Iron losses}$$

Calculations : We know that,

$$W_o = V_o I_o \cos \phi$$

$$\therefore \cos \phi_o = \frac{W_o}{V_o I_o} = \text{No load power factor}$$

Once $\cos \phi_o$ is known we can obtain,

$$I_c = I_o \cos \phi_o$$

$$I_m = I_o \sin \phi_o$$

and

Once I_c and I_m are known we can determine exciting circuit parameters as,

$$R_o = \frac{V_o}{I_c} \quad \Omega$$

$$X_o = \frac{V_o}{I_m} \quad \Omega$$

and

Key Point : The no load power factor $\cos \phi_o$ is very low hence wattmeter used must be low power factor type otherwise there might be error in the results. If the meters are connected on secondary and primary is kept open then from O.C. test we get R'_o and X'_o with which we can obtain R_o and X_o knowing the transformation ratio K .

Short Circuit Test (S.C. Test)

In this test, primary is connected to a.c. supply through variac, ammeter and voltmeter as shown in the Fig. 3.27.

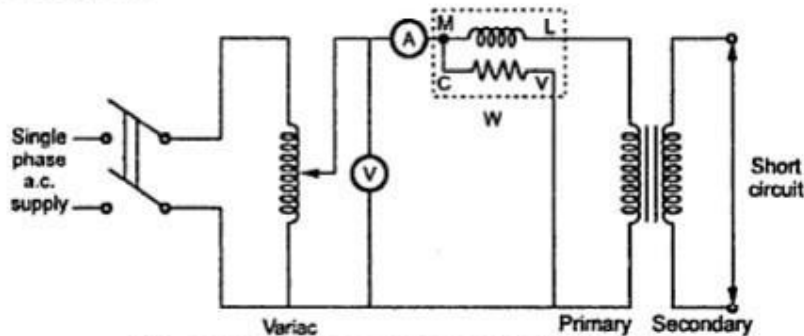


Fig. 3.27 Experimental circuit for S.C. test

The secondary is short circuited with the help of thick copper wire or solid link. As high voltage side is always low current side, it is convenient to connect high voltage side to supply and shorting the low voltage side.

BEEE (Part-A)

As secondary is shorted, its resistance is very very small and on rated voltage it may draw very large current. Such large current can cause overheating and burning of the transformer. To limit this short circuit current, primary is supplied with low voltage which is just enough to cause rated current to flow through primary which can be observed on an ammeter. The low voltage can be adjusted with the help of variac. Hence this test is also called low voltage test or reduced voltage test. The wattmeter reading as well as voltmeter, ammeter readings are recorded. The observation table is as follows.

V_{sc} volts	I_{sc} amperes	W_{sc} watts
	Rated	

Now the currents flowing through the windings are rated currents hence the total copper loss is full load copper loss. Now the voltage applied is low which is a small fraction of the rated voltage. The iron losses are function of applied voltage. So the iron losses in reduced voltage test are very small. Hence the wattmeter reading is the power loss which is equal to full load copper losses as iron losses are very low.

$$\therefore W_{sc} = (P_{cu}) \text{ F.L.} = \text{Full load copper loss}$$

Calculations : From S.C. test readings we can write,

$$W_{sc} = V_{sc} I_{sc} \cos \phi_{sc}$$

$$\therefore \cos \phi_{sc} = \frac{V_{sc} I_{sc}}{W_{sc}} = \text{Short circuit power factor}$$

$$W_{sc} = I_{sc}^2 R_{1e} = \text{Copper loss}$$

$$\therefore R_{1e} = \frac{W_{sc}}{I_{sc}^2}$$

While $Z_{1e} = \frac{V_{sc}}{I_{sc}} = \sqrt{R_{1e}^2 + X_{1e}^2}$

$$\therefore X_{1e} = \sqrt{Z_{1e}^2 - R_{1e}^2}$$

Thus we get the equivalent circuit parameters R_{1e} , X_{1e} and Z_{1e} . Knowing the transformation ratio K , the equivalent circuit parameters referred to secondary also can be obtained.

calculated from S.C. test readings are referred to secondary which are R_{2e} , Z_{2e} and X_{2e} . So before doing calculations it is necessary to find out where the readings are recorded on transformer primary or secondary and accordingly the parameters are to be determined. In step down transformer, primary is high voltage itself to which supply is given in S.C. test. So in such case test results give us parameters referred to primary i.e. R_{1e} , Z_{1e} and X_{1e} .

Losses in a Transformer

In a transformer, there exists two types of losses.

- i) The core gets subjected to an alternating flux, causing **core losses**.
- ii) The windings carry currents when transformer is loaded, causing **copper losses**.

Core or Iron Losses

Due to alternating flux set up in the magnetic core of the transformer, it undergoes a cycle of magnetization and demagnetization. Due to hysteresis effect there is loss of energy in this process which is called hysteresis loss.

It is given by, hysteresis loss = $K_h B_m^{1.67} f v$ watts

where K_h = Hysteresis constant depends on material

B_m = Maximum flux density

f = Frequency

v = Volume of the core

The induced e.m.f. in the core tries to set up eddy currents in the core and hence responsible for the eddy current losses. The eddy current loss is given by,

Eddy current loss = $K_e B_m^2 f^2 t^2$ watts/unit volume

where K_e = Eddy current constant

t = Thickness of the core

As seen earlier, the flux in the core is almost constant as supply voltage V_1 at rated frequency f is always constant. Hence the flux density B_m in the core and hence both hysteresis and eddy current losses are constants at all the loads. Hence the core or iron losses are also called **constant losses**. The iron losses are denoted as P_i .

The iron losses are minimized by using high grade core material like silicon steel having very low hysteresis loop and by manufacturing the core in the form of laminations.

Copper Losses

The copper losses are due to the power wasted in the form of I^2R loss due to the resistances of the primary and secondary windings. The copper loss depends on the magnitude of the currents flowing through the windings.

$$\begin{aligned} \text{Total Cu loss} &= I_1^2 R_1 + I_2^2 R_2 = I_1^2 (R_1 + R_2') = I_2^2 (R_2 + R_1') \\ &= I_1^2 R_{1e} = I_2^2 R_{2e} \end{aligned}$$

The copper losses are denoted as P_{cu} . If the current through the windings is full load current, we get copper losses at full load. If the load on transformer is half then we get copper losses at half load which are less than full load copper losses. Thus copper losses are called **variable losses**. For transformer VA rating is $V_1 I_1$ or $V_2 I_2$. As V_1 is constant, we can say that copper losses are proportional to the square of the kVA rating.

So, $P_{cu} \propto I^2 \propto (\text{kVA})^2$

Thus for a transformer,

$$\begin{aligned} \text{Total losses} &= \text{Iron losses} + \text{Copper losses} \\ &= P_i + P_{cu} \end{aligned}$$

E.M.F. Equation of a Transformer

When the primary winding is excited by an alternating voltage V_1 , it circulates alternating current, producing an alternating flux ϕ . The primary winding has N_1 number of turns. The alternating flux ϕ linking with the primary winding itself induces an e.m.f. in it denoted as E_1 . The flux links with secondary winding through the common magnetic core. It produces induced e.m.f. E_2 in the secondary winding. This is mutually induced e.m.f. Let us derive the equations for E_1 and E_2 .

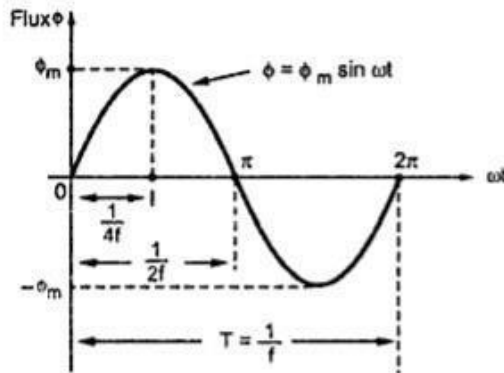


Fig. 3.11 Sinusoidal flux

The primary winding is excited by purely sinusoidal alternating voltage. Hence the flux produced is also sinusoidal in nature having maximum value of ϕ_m as shown in the Fig. 3.11.

The various quantities which affect the magnitude of the induced e.m.f. are :

ϕ = Flux

ϕ_m = Maximum value of flux

N_1 = Number of primary winding turns

N_2 = Number of secondary winding turns

f = Frequency of the supply voltage

E_1 = R.M.S. value of the primary induced e.m.f.

E_2 = R.M.S. value of the secondary induced e.m.f.

From Faraday's law of electromagnetic induction the average e.m.f. induced in each turn is proportional to the average rate of change of flux.

\therefore Average e.m.f. per turn = Average rate of change of flux

\therefore Average e.m.f. per turn = $\frac{d\phi}{dt}$

Now $\frac{d\phi}{dt} = \frac{\text{Change in flux}}{\text{Time required for change in flux}}$

Consider the $1/4^{\text{th}}$ cycle of the flux as shown in the Fig. 3.11. Complete cycle gets completed in $1/f$ seconds. In $1/4^{\text{th}}$ time period, the change in flux is from 0 to ϕ_m .

$\therefore \frac{d\phi}{dt} = \frac{\phi_m - 0}{\left(\frac{1}{4f}\right)}$ as dt for $1/4^{\text{th}}$ time period is $1/4f$ seconds

BEEE (Part-A)

$$= 4 f \phi_m \text{ Wb/sec}$$

$$\therefore \text{Average e.m.f. per turn} = 4 f \phi_m \text{ volts}$$

As ϕ is sinusoidal, the induced e.m.f. in each turn of both the windings is also sinusoidal in nature. For sinusoidal quantity,

$$\text{Form factor} = \frac{\text{R.M.S. value}}{\text{Average value}} = 1.11$$

$$\therefore \text{R.M.S. value} = 1.11 \times \text{Average value}$$

$$\begin{aligned} \therefore \text{R.M.S. value of induced e.m.f. per turn} \\ &= 1.11 \times 4 f \phi_m \\ &= 4.44 f \phi_m \end{aligned}$$

There are N_1 number of primary turns hence the R.M.S. value of induced e.m.f. of primary denoted as E_1 is,

$$E_1 = N_1 \times 4.44 f \phi_m \text{ volts}$$

While as there are N_2 number of secondary turns the R.M.S value of induced e.m.f. of secondary denoted E_2 is,

$$E_2 = N_2 \times 4.44 f \phi_m \text{ volts}$$

The expressions of E_1 and E_2 are called e.m.f. equations of a transformer.

Thus e.m.f. equations are,

$$E_1 = 4.44 f \phi_m N_1 \text{ volts} \quad \dots (1)$$

$$E_2 = 4.44 f \phi_m N_2 \text{ volts} \quad \dots (2)$$

Voltage Ratio

We know from the e.m.f. equations of a transformer that

$$E_1 = 4.44 f \phi_m N_1 \quad \text{and} \quad E_2 = 4.44 f \phi_m N_2$$

Taking ratio of the two equations we get,

$$\frac{E_2}{E_1} = \frac{N_2}{N_1} = K$$

This ratio of secondary induced e.m.f. to primary induced e.m.f. is known as voltage transformation ratio denoted as K .

Thus,

$$E_2 = K E_1 \quad \text{where} \quad K = \frac{N_2}{N_1}$$

1. If $N_2 > N_1$ i.e. $K > 1$, we get $E_2 > E_1$ then the transformer is called **step-up transformer**.
2. If $N_2 < N_1$ i.e. $K < 1$, we get $E_2 < E_1$ then the transformer is called **step-down transformer**.
3. If $N_2 = N_1$ i.e. $K = 1$, we get $E_2 = E_1$ then the transformer is called **isolation transformer or 1:1 transformer**.

Current Ratio

For an ideal transformer there are no losses. Hence the product of primary voltage V_1 and primary current I_1 , is same as the product of secondary voltage V_2 and the secondary current I_2 .

$$\text{So } V_1 I_1 = \text{input VA} \quad \text{and} \quad V_2 I_2 = \text{output VA}$$

For an ideal transformer,

$$V_1 I_1 = V_2 I_2$$

∴

$$\boxed{\frac{V_2}{V_1} = \frac{I_1}{I_2} = K}$$

Key Point: Hence the currents are in the inverse ratio of the voltage transformation ratio.

Volt-Ampere Rating

When electrical power is transferred from primary winding to secondary there are few power losses in between. These power losses appear in the form of heat which increase the temperature of the device. Now this temperature must be maintained below certain limiting value as it is always harmful from insulation point of view. As current is the main cause in producing heat, the output maximum rating is generally specified as the product of output voltage and output current i.e. $V_2 I_2$. This always indicates that when transformer is operated under this specified rating, its temperature rise will not be excessive. The copper losses depend on current and iron losses depend on voltage. These losses are independent of the load power factor $\cos \phi_2$. Hence though the output power depends on $\cos \phi_2$, the transformer losses are functions of V and I and the rating of the transformer is specified as the product of voltage and current called VA rating. This rating is generally expressed in kVA (kilo volt amperes rating).

$$\text{Now } \frac{V_1}{V_2} = \frac{I_2}{I_1} = K$$

$$\therefore V_1 I_1 = V_2 I_2$$

$$\boxed{\text{kVA rating of a transformer} = \frac{V_1 I_1}{1000} = \frac{V_2 I_2}{1000}}$$

If V_1 and V_2 are the terminal voltages of primary and secondary then from specified kVA rating we can decide full load currents of primary and secondary, I_1 and I_2 . This is the safe maximum current limit which may carry, keeping temperature rise below its limiting value.

$$\boxed{\begin{aligned} I_1 \text{ full load} &= \frac{\text{kVA rating} \times 1000}{V_1} \quad \dots \text{ (1000 to convert kVA to VA)} \\ I_2 \text{ full load} &= \frac{\text{kVA rating} \times 1000}{V_2} \end{aligned}}$$

Key Point: The full load primary and secondary currents indicate the safe maximum values of currents which transformer windings can carry.

These values indicate, how much maximum load can be connected to a given transformer of a specified kVA rating.

All Day Efficiency of a Transformer

For a transformer, the efficiency is defined as the ratio of output power to input power. This is its power efficiency. But power efficiency is not the true measure of the performance of some special types of transformers such as distribution transformers.

Distribution transformers serve residential and commercial loads. The load on such transformers vary considerably during the period of the day. For most period of the day these transformers are working at 30 to 40 % of full load only or even less than that. But the primary of such transformers is energised at its rated voltage for 24 hours, to provide continuous supply to the consumer. The core loss which depends on voltage, takes place continuously for all the loads. But copper loss depends on the load condition. For no load, copper loss is negligibly small while on full load it is at its rated value. Hence power efficiency cannot give the measure of true efficiency of such transformers. In such transformers, the energy output is calculated in kilo watt hours (kWh). Similarly energy spent in supplying the various losses is also determined in kilo watt hours (kWh). Then ratio of total energy output to total energy input (output + losses) is calculated. Such ratio is called **Energy efficiency** or **All Day Efficiency** of a transformer. Based on this efficiency, the performance of various distribution transformers is compared. All day efficiency is defined as,

$$\begin{aligned} \% \text{ All day } \eta &= \frac{\text{Output energy in kWh during a day}}{\text{Input energy in kWh during a day}} \times 100 \\ &= \frac{\text{Output energy in kWh during a day}}{\text{Output energy} + \text{Energy spent for total losses}} \times 100 \end{aligned}$$

While calculating energies, all energies can be expressed in watt hour (Wh) instead of kilo watt hour (kWh).

Such distribution transformers are designed to have very low core losses. This is achieved by limiting the core flux density to lower value by using a relatively higher core cross-section i.e. larger iron to copper weight ratio. The maximum efficiency in such transformers occur at about 60-70 % of the full load. So by proper designing, high energy efficiencies can be achieved for distribution transformers.

PRINCIPLE AND OPERATION OF INDUCTION MOTOR

Introduction

An electric motor is a device which converts an electrical energy into a mechanical energy. This mechanical energy then can be supplied to various types of loads. The motors can operate on d.c. as well as single and three phase a.c. supply. The motors operating on d.c. supply are called d.c. motors while motors operating on a.c. supply are called a.c. motors. As a.c. supply is commonly available, the a.c. motors are very popularly used in practice. The a.c. motors are classified as single and three phase induction motors, synchronous motors and some special purpose motors. Out of all these types, three phase induction motors are widely used for various industrial applications. Hence this chapter gives the emphasis on the working principle, types and features of three phase induction motors. The important advantages of three phase induction motors over other types are self starting property, no need of starting device, higher power factor, good speed regulation and robust construction. The working principle of three phase induction motors is based on the production of **rotating magnetic field**.

Construction

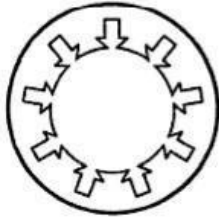
Basically, the induction motor consists of two main parts, namely

1. The part i.e. three phase windings, which is stationary called stator.
2. The part which rotates and is connected to the mechanical load through shaft called rotor.

The conversion of electrical power to mechanical power takes place in a rotor. Hence rotor develops a driving torque and rotates.

Stator

The stator has a laminated type of construction made up of stampings which are 0.4 to 0.5 mm thick. The stampings are slotted on its periphery to carry the stator winding. The stampings are insulated from each other. Such a construction essentially keeps the iron losses to a minimum value. The number of stampings are stamped together to build the stator core. The built up core is then fitted in a casted or fabricated steel frame. The choice of material for the stampings is generally silicon steel, which minimises the hysteresis loss. The slots on the periphery of the stator core carries a three phase winding, connected either in star or delta. This three phase winding is called stator winding. It is wound for definite number of poles. This winding when excited by a three phase supply produces a rotating magnetic field as discussed earlier. The choice of number of poles depends on the speed of the rotating magnetic field required.



Stator lamination

The radial ducts are provided for the cooling purpose. In some cases, all the six terminals of three phase stator winding are brought out which gives flexibility to the user to connect them either in star or delta. The Fig. 5.6 shows a stator lamination.

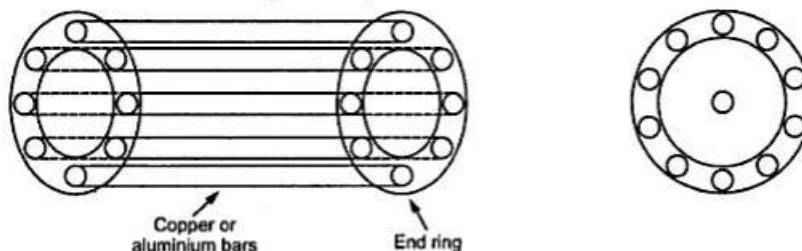
Rotor

The rotor is placed inside the stator. The rotor core is also laminated in construction and uses cast iron. It is cylindrical, with slots on its periphery. The rotor conductors or winding is placed in the rotor slots. The two types of rotor constructions which are used for induction motors are,

1. Squirrel cage rotor and
2. Slip ring or wound rotor

Squirrel Cage Rotor

The rotor core is cylindrical and slotted on its periphery. The rotor consists of uninsulated copper or aluminium bars called rotor conductors. The bars are placed in the slots. These bars are permanently shorted at each end with the help of conducting copper ring called end ring. The bars are usually brazed to the end rings to provide good mechanical strength. The entire structure looks like a cage, forming a closed electrical circuit. So the rotor is called squirrel cage rotor. The construction is shown in the Fig. 5.7.



(a) Cage type structure of rotor

(b) Symbolic representation

Fig. 5.7 Squirrel cage rotor

As the bars are permanently shorted to each other through end ring, the entire rotor resistance is very very small. Hence this rotor is also called short circuited rotor. As rotor itself is short circuited, no external resistance can have any effect on the rotor resistance. Hence no external resistance can be introduced in the rotor circuit. So slip ring and brush assembly is not required for this rotor. Hence the construction of this rotor is very simple.

Fan blades are generally provided at the ends of the rotor core. This circulates the air through the machine while operation, providing the necessary cooling. The air gap between stator and rotor is kept uniform and as small as possible.

BEEE (Part-A)

$$N_s = \frac{120 f}{P} = \text{Speed of rotating magnetic field.}$$

where

f = Supply frequency.

P = Number of poles for which stator winding is wound.

This rotating field produces an effect of rotating poles around a rotor. Let direction of rotation of this rotating magnetic field is clockwise as shown in the Fig. 5.9 (a).

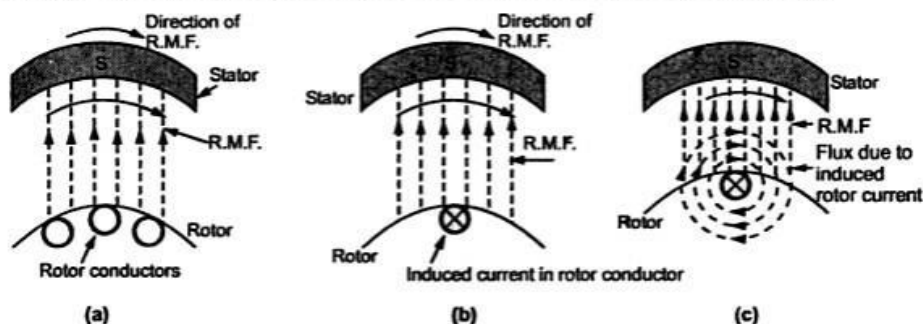


Fig. 5.9

Now at this instant rotor is stationary and stator flux R.M.F. is rotating. So its obvious that there exists a relative motion between the R.M.F. and rotor conductors. Now the R.M.F. gets cut by rotor conductors as R.M.F. sweeps over rotor conductors. Whenever conductor cuts the flux, e.m.f. gets induced in it. So e.m.f. gets induced in the rotor conductors called rotor induced e.m.f. This is electro-magnetic induction. As rotor forms closed circuit, induced e.m.f. circulates current through rotor called rotor current as shown in the Fig. 5.9 (b). Let direction of this current is going into the paper denoted by a cross as shown in the Fig. 5.9 (b).

Any current carrying conductor produces its own flux. So rotor produces its flux called rotor flux. For assumed direction of rotor current, the direction of rotor flux is clockwise as shown in the Fig. 5.9 (c). This direction can be easily determined using right hand thumb rule. Now there are two fluxes, one R.M.F. and other rotor flux. Both the fluxes interact with each as shown in the Fig. 5.9 (d). On left of rotor conductor, two fluxes are in same direction hence add up to get high flux area. On right side, two fluxes cancel each other to produce low flux area. As flux lines act as stretched rubber band, high flux density area exerts a push on rotor conductor towards low flux density area. So rotor conductor experiences a force from left to right in this case, as shown in the Fig. 5.9 (d), due to interaction of the two fluxes.

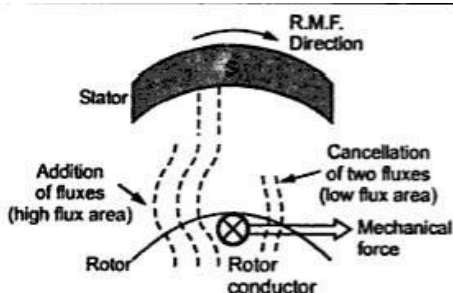


Fig. 5.9 (d)

As all the rotor conductors experience a force, the overall rotor experiences a torque and starts rotating. So interaction of the two fluxes is very essential for a motoring action. As seen from the Fig. 5.9 (d), the direction of force experienced is same as that of rotating magnetic field. Hence rotor starts rotating in the same direction as that of rotating magnetic field.

Alternatively this can be explained as : According to Lenz's law the direction of induced current in the rotor is so as to oppose the cause producing it. The cause of rotor current is the induced e.m.f. which is induced because of relative motion present between the rotating magnetic field and the rotor conductors. Hence to oppose the relative motion i.e. to reduce the relative speed, the rotor experiences a torque in the same direction as that of R.M.F. and tries to catch up the speed of rotating magnetic field.

So, N_s = Speed of rotating magnetic field in r.p.m.

N = Speed of rotor i.e. motor in r.p.m.

$N_s - N$ = Relative speed between the two, rotating magnetic field and the rotor conductors.

Thus rotor always rotates in same direction as that of R.M.F.

Can $N = N_s$?

When rotor starts rotating, it tries to catch the speed of rotating magnetic field.

If it catches the speed of the rotating magnetic field, the relative motion between rotor and the rotating magnetic field will vanish ($N_s - N = 0$). In fact the relative motion is the main cause for the induced e.m.f. in the rotor. So induced e.m.f. will vanish and hence there cannot be rotor current and the rotor flux which is essential to produce the torque on the rotor. Eventually motor will stop. But immediately there will exist a relative motion between rotor and rotating magnetic field and it will start. But due to inertia of rotor, this does not happen in practice and rotor continues to rotate with a speed slightly less than the synchronous speed of the rotating magnetic field in the steady state. The induction motor never rotates at synchronous speed. The speed at which it rotates is hence called subsynchronous speed and motor sometimes called asynchronous motor.

$$\therefore N < N_s$$

So it can be said that rotor slips behind the rotating magnetic field produced by stator. The difference between the two is called slip speed of the motor.

$$N_s - N = \text{Slip speed of the motor in r.p.m.}$$

This speed decides the magnitude of the induced e.m.f. and the rotor current, which in turn decides the torque produced. The torque produced is as per the requirements of overcoming the friction and iron losses of the motor along with the torque demanded by the load on the motor.

Slip of Induction Motor

We have seen that rotor rotates in the same direction as that of R.M.F. but in steady state attains a speed less than the synchronous speed. The difference between the two speeds i.e. synchronous speed of R.M.F. (N_s) and rotor speed (N) is called slip speed. This slip speed is generally expressed as the percentage of the synchronous speed.

So slip of the induction motor is defined as the difference between the synchronous speed (N_s) and actual speed of rotor i.e. motor (N) expressed as a fraction of the synchronous speed (N_s). This is also called absolute slip or fractional slip and is denoted as 's'.

$$\text{Thus } s = \frac{N_s - N}{N_s} \quad \dots \text{ (Absolute slip)}$$

The percentage slip is expressed as,

$$\% s = \frac{N_s - N}{N_s} \times 100 \quad \dots \text{ (Percentage slip)}$$

In terms of slip, the actual speed of motor (N) can be expressed as,

$$N = N_s (1 - s) \quad \dots \text{ (From the expression of slip)}$$

At start, motor is at rest and hence its speed N is zero.

$$\therefore s = 1 \text{ at start}$$

This is maximum value of slip s possible for induction motor which occurs at start. While $s = 0$ gives us $N = N_s$ which is not possible for an induction motor. So slip of induction motor cannot be zero under any circumstances.

Practically motor operates in the slip range of 0.01 to 0.05 i.e. 1 % to 5 %. The slip corresponding to full load speed of the motor is called full load slip.

BEEE (Part-A)

► **Example :** A 4 pole, 3 phase induction motor is supplied from 50 Hz supply. Determine its synchronous speed. On full load, its speed is observed to be 1410 r.p.m. Calculate its full load slip.

Solution : Given values are,

$$P = 4, \quad f = 50 \text{ Hz}, \quad N = 1410 \text{ r.p.m.}$$

$$N_s = \frac{120 f}{P} = \frac{120 \times 50}{4} = 1500 \text{ r.p.m.}$$

Full load absolute slip is given by,

$$s = \frac{N_s - N}{N_s} = \frac{1500 - 1410}{1500} = 0.06$$

$$\therefore \% s = 0.06 \times 100 = 6 \%$$

► **Example :** A 4 pole, 3 phase, 50 Hz, star connected induction motor has a full load slip of 4 %. Calculate full load speed of the motor.

Solution : Given values are,

$$P = 4, \quad f = 50 \text{ Hz}, \quad \% s_{fl} = 4 \%$$

$$s_{fl} = \text{Full load absolute slip} = 0.04$$

$$N_s = \frac{120 f}{P} = \frac{120 \times 50}{4} = 1500 \text{ r.p.m.}$$

$$s_{fl} = \frac{N_s - N_{fl}}{N_s} \quad \text{where } N_{fl} = \text{Full load speed of motor}$$

$$\therefore 0.04 = \frac{1500 - N_{fl}}{1500}$$

$$\therefore N_{fl} = 1440 \text{ r.p.m.}$$

This is the full load speed of the motor.

UNIT – 3 BASIC OF POWER SYSTEMS

Sources of Energy

Since electrical energy is produced from energy available in various forms in nature, it is desirable to look into the various sources of energy. These sources of energy are :

- (i) The Sun (ii) The Wind (iii) Water (iv) Fuels (v) Nuclear energy.

Out of these sources, the energy due to Sun and wind has not been utilised on large scale due to a number of limitations. At present, the other three sources *viz.*, water, fuels and nuclear energy are primarily used for the generation of electrical energy.

(i) **The Sun.** The Sun is the primary source of energy. The heat energy radiated by the Sun can be focussed over a small area by means of reflectors. This heat can be used to raise steam and electrical energy can be produced with the help of turbine-alternator combination. However, this method has limited application because :

- (a) it requires a large area for the generation of even a small amount of electric power
- (b) it cannot be used in cloudy days or at night
- (c) it is an uneconomical method.

Nevertheless, there are some locations in the world where strong solar radiation is received very regularly and the sources of mineral fuel are scanty or lacking. Such locations offer more interest to the solar plant builders.

(i) **The Sun.** The Sun is the primary source of energy. The heat energy radiated by the Sun can be focussed over a small area by means of reflectors. This heat can be used to raise steam and electrical energy can be produced with the help of turbine-alternator combination. However, this method has limited application because :

- (a) it requires a large area for the generation of even a small amount of electric power
- (b) it cannot be used in cloudy days or at night
- (c) it is an uneconomical method.

Nevertheless, there are some locations in the world where strong solar radiation is received very regularly and the sources of mineral fuel are scanty or lacking. Such locations offer more interest to the solar plant builders.

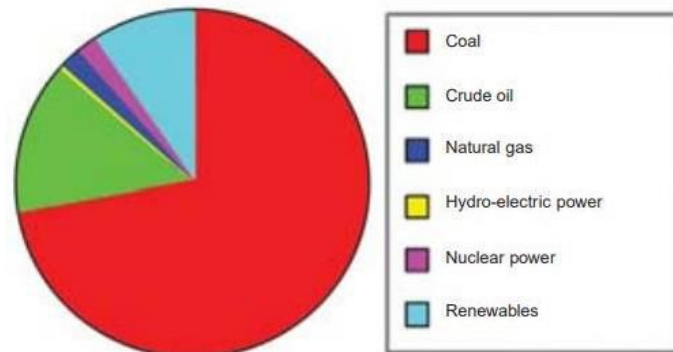
(ii) **The Wind.** This method can be used where wind flows for a considerable length of time. The wind energy is used to run the wind mill which drives a small generator. In order to obtain the electrical energy from a wind mill continuously, the generator is arranged to charge the batteries. These batteries supply the energy when the wind stops. This method has the advantages that maintenance and generation costs are negligible. However, the drawbacks of this method are (a) variable output, (b) unreliable because of uncertainty about wind pressure and (c) power generated is quite small.

(iii) **Water.** When water is stored at a suitable place, it possesses potential energy because of the head created. This water energy can be converted into mechanical energy with the help of water turbines. The water turbine drives the alternator which converts mechanical energy into electrical energy. This method of generation of electrical energy has become very popular because it has low production and maintenance costs.

(iv) **Fuels.** The main sources of energy are fuels *viz.*, solid fuel as coal, liquid fuel as oil and gas fuel as natural gas. The heat energy of these fuels is converted into mechanical energy by suitable prime movers such as steam engines, steam turbines, internal combustion engines etc. The prime mover drives the alternator which converts mechanical energy into electrical energy. Although fuels continue to enjoy the place of chief source for the generation of electrical energy, yet their reserves are diminishing day by day. Therefore, the present trend is to harness water power which is more or less a permanent source of power.

(v) **Nuclear energy.** Towards the end of Second World War, it was discovered that large amount of heat energy is liberated by the *fission* of uranium and other fissionable materials. It is estimated that heat produced by 1 kg of nuclear fuel is equal to that produced by 4500 tonnes of coal. The heat produced due to nuclear fission can be utilised to raise steam with suitable arrangements. The steam

can run the steam turbine which in turn can drive the alternator to produce electrical energy. However, there are some difficulties in the use of nuclear energy. The principal ones are (a) high cost of nuclear plant (b) problem of disposal of radioactive waste and dearth of trained personnel to handle the plant.



Energy Utilisation

LAYOUT AND OPERATION OF HYDRO ELECTRIC POWER STATION

Hydro-electric Power Station

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 available 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 service instantly.
- (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

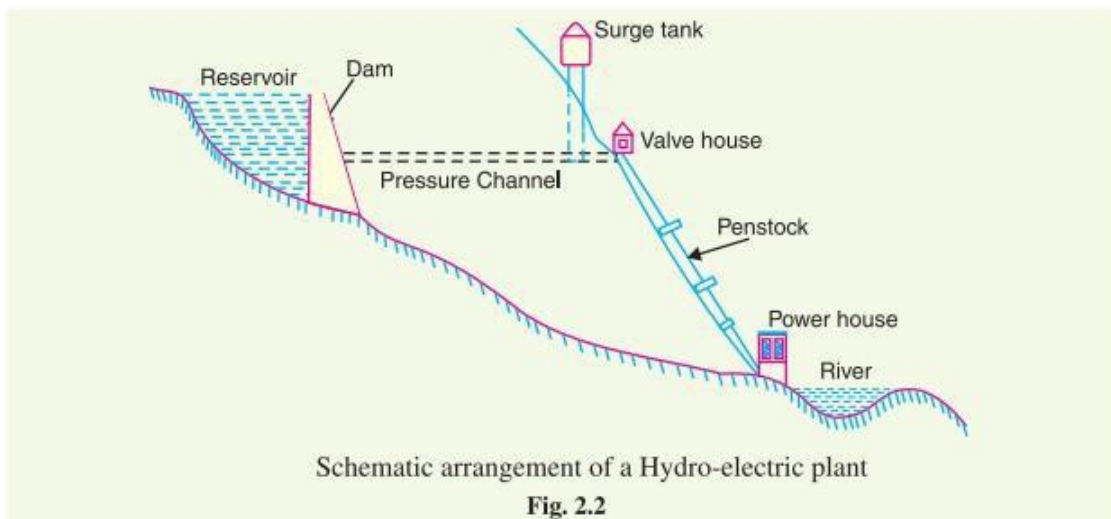
- (i) It involves high capital cost due to construction of dam.
- (ii) There is uncertainty about the availability of huge amount of water due to dependence on weather conditions.
- (iii) Skilled and experienced hands are required to build the plant.
- (iv) It requires high cost of transmission lines as the plant is located in hilly areas which are quite away from the consumers.

Schematic Arrangement of Hydro-electric 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.



LAYOUT AND OPERATION OF THERMAL / 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 utilising 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

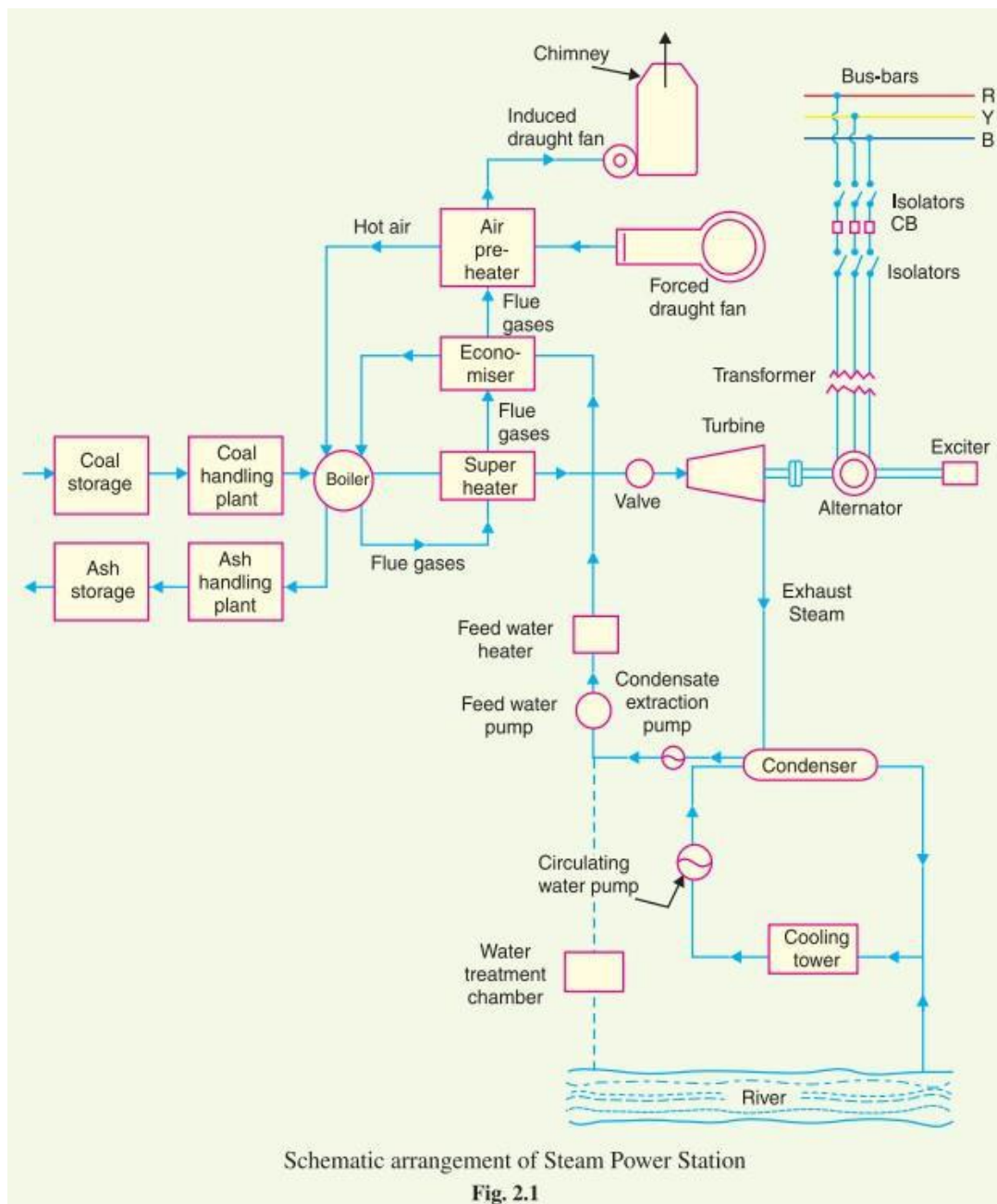
- (i) It pollutes the atmosphere due to the production of large amount of smoke and fumes.
- (ii) It is costlier in running cost as compared to hydroelectric plant.

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. 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

1. 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.

2. 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 super-heater, 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.

3. 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.

4. 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.

5. 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.

6. 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.

LAYOUT AND OPERATION OF NUCLEAR POWER STATION

Nuclear Power Station

A generating station in which nuclear energy is converted into electrical energy is known as a **nuclear power 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 electrical energy.

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 in the 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 electric power.
- (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 is reduced.
- (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.

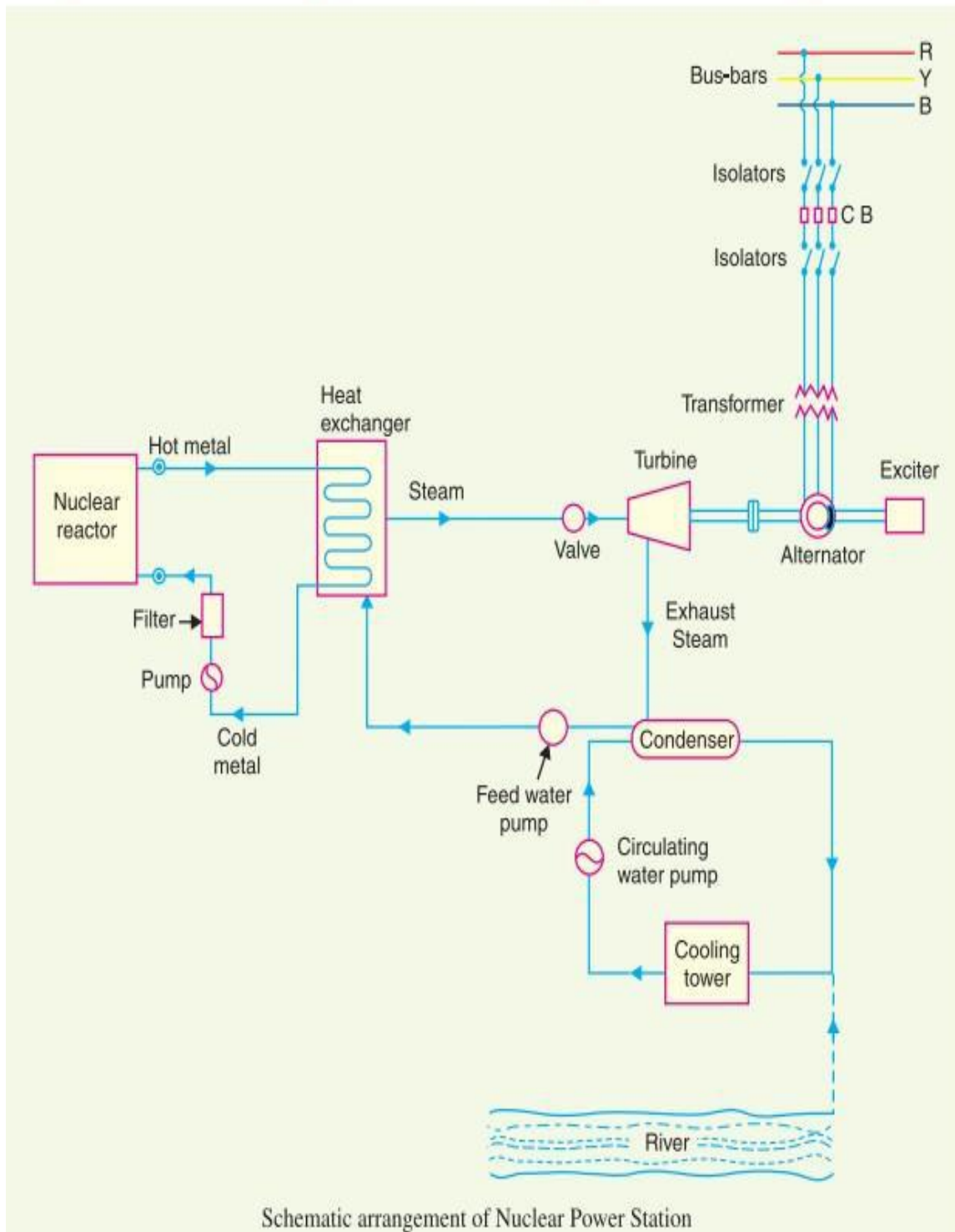
Disadvantages

- (i) The fuel used is expensive and is difficult to recover.
- (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 standardisation. 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 fluctuations efficiently.
- (vii) The disposal of the by-products, which are radioactive, is a big problem. They have either to be disposed off in a deep trench or in a sea away from sea-shore.

Schematic Arrangement of Nuclear Power Station

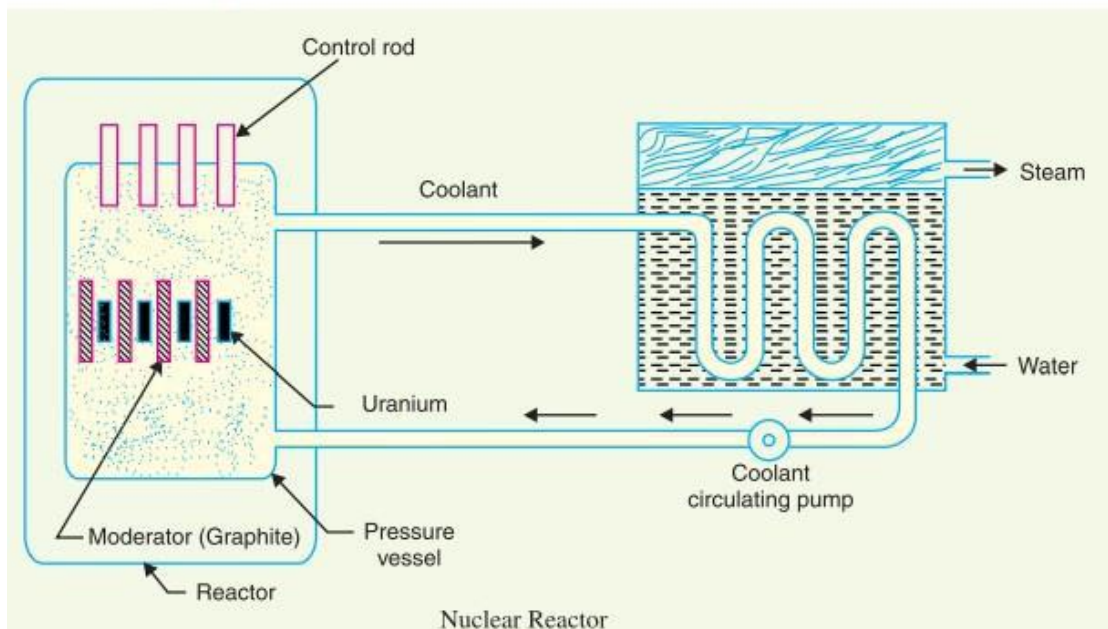
The schematic arrangement of a nuclear power station is shown in Fig. The whole arrangement can be divided into the following main stages :

- (i) Nuclear reactor (ii) Heat exchanger (iii) Steam turbine (iv) Alternator.



- (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 by pushing 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 the reactor is removed by the coolant, generally a sodium metal. The coolant carries the heat to the heat exchanger.

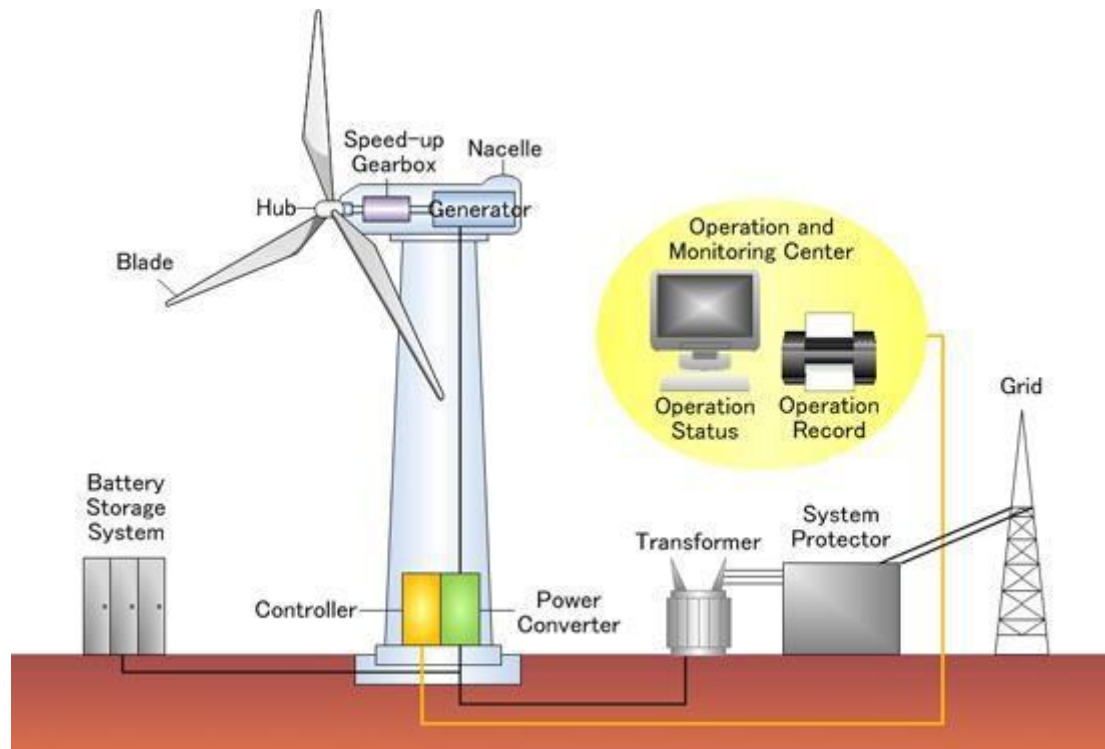


- (ii) **Heat exchanger.** The coolant gives up heat to the heat exchanger which is utilised in raising the steam. After giving up heat, the coolant is again fed to the reactor.

* **Chain reaction.** Nuclear fission is done by bombarding Uranium nuclei with slow moving neutrons. This splits the Uranium nuclei with the release of huge amount of energy and emission of neutrons (called fission neutrons). These fission neutrons cause further fission. If this process continues, then in a very short time huge amount of energy will be released which may cause explosion. This is known as *explosive chain reaction*. But in a reactor, controlled chain reaction is allowed. This is done by systematically removing the fission neutrons from the reactor. The greater the number of fission neutrons removed, the lesser is the intensity (*i.e.*, fission rate) of energy released.

- (iii) **Steam turbine.** The steam produced in the heat exchanger 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 transformer, circuit breakers and isolators.

LAYOUT AND OPERATION OF WIND POWER PLANT



There is an air turbine of large blades attached on the top of a supporting tower of sufficient height. When wind strikes on the turbine blades, the turbine rotates due to the design and alignment of rotor blades. The shaft of the turbine is coupled with an electrical generator. The output of the generator is collected through electric power cables.

Working of Wind Turbine

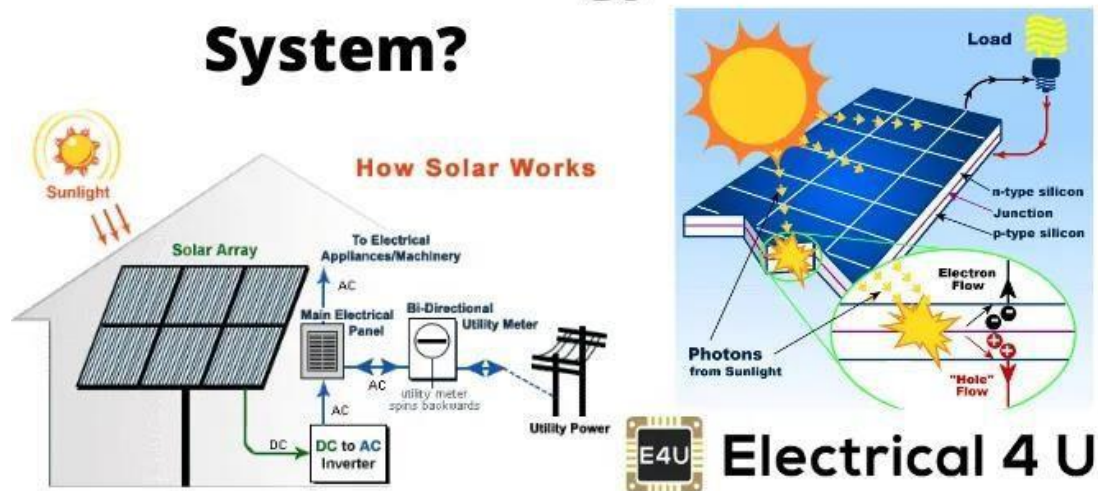
When the wind strikes the rotor blades, blades start rotating. The turbine rotor is connected to a high-speed gearbox. Gearbox transforms the rotor rotation from low speed to high speed. The high-speed shaft from the gearbox is coupled with the rotor of the generator and hence the electrical generator runs at a higher speed. An exciter is needed to give the required excitation to the magnetic coil of the generator field system so that it can generate the required electricity. The generated voltage at output terminals of the alternator is proportional to both the speed and field flux of the alternator.

The speed is governed by wind power which is out of control. Hence to maintain uniformity of the output power from the alternator, excitation must be controlled according to the availability of natural wind power. The exciter current is controlled by a turbine controller which senses the wind speed. Then output voltage of electrical generator(alternator) is given to a rectifier where the alternator output gets rectified to DC. Then this rectified DC output is given to line converter unit to convert it into stabilized AC output which is ultimately fed to either electrical transmission network or transmission grid with the help of step up transformer. An extra units is used to give the power to internal

auxiliaries of **wind turbine** (like motor, battery etc.), this is called Internal Supply Unit.

LAYOUT AND OPERATION OF SOLAR POWER PLANT

What is Solar Energy System?



Solar Energy System

Solar energy is the light and radiant heat from the Sun that control Earth's climate and weather and protract life. It is a renewable source of energy and originates with the thermonuclear process that transfers about 650,000,000 tons of hydrogen to helium per second. This action produces lots of heat and electromagnetic radiation. The produced heat remains in the sun and is helpful in upholding the thermonuclear reaction and electromagnetic radiation together with visible, infrared and ultra-violet radiation flow out into space in all directions.

Solar energy system is the pollution free source of energy and always available because, sun is the single source of solar energy (also known as renewable energy or non-conventional energy) which sits at the central point of solar system and radiate energy at an tremendously huge and fairly constant rate, per day per year as the form of electromagnetic radiation. Sun contained huge amount of energy.

Working Principle of Solar Power Plant:

The solar power plant is also known as the Photovoltaic (PV) power plant. It is a large-scale PV plant designed to produce bulk electrical power from solar radiation. The solar power plant uses solar energy to produce electrical power. Therefore, it is a conventional power plant.

Solar energy can be used directly to produce electrical energy using solar PV panels. Or there is another way to produce electrical energy that is concentrated solar energy. In this type of plant, the radiation energy of solar first converted into heat (thermal energy) and this heat is used to drive a conventional generator. This method is difficult and not efficient to produce electrical power on a large scale. Hence, to produce electrical power on a large scale, solar PV panels are used.

Photo Voltaic (PV) Principle

Silicon is the most commonly used material in solar cells. Silicon is a semiconductor material. Several materials show photoelectric properties like; cadmium, gallium arsenide, etc.

Electron-holes pairs are created in solar cells. The PV materials have the property to absorb photons of sunlight. The valance band electrons of semiconductor material are at lower energy and the electrons of conduction band are at a higher energy level. The difference between this energy level is known as band gap energy E_g .

When sunlight falls on solar cells, the difference between photon energy E and bandgap energy E_g is absorbed by the cell. And it excites some electrons to jump across the bandgap. These electrons move from the valance band to the conduction band and create holes in the valance band.

Components of Solar Power Plant

The major components of the solar photovoltaic system are listed below.

- Photovoltaic (PV) panel
- Inverter
- Energy storage devices
- Charge controller

- **Photovoltaic (PV) Panel**

PV panels or Photovoltaic panel is a most important component of a solar power plant. It is made up of small solar cells. This is a device that is used to convert solar photon energy into electrical energy.

Generally, silicon is used as a semiconductor material in solar cells. The typical rating of silicon solar cells is 0.5 V and 6 Amp. And it is equivalent to 3 W power. The number of [cells is connected in series or parallel](#) and makes a module. The number of modules forms a solar panel.

According to the capacity of power plants, a number of plates are mounted and a group of panels is also known as [Photovoltaic \(PV\) array](#).

- **Inverter**
- The output of the solar panel is in the form of DC. The most of load connected to the [power system](#) network is in the form of AC. Therefore, we need to convert DC output power into AC power. For that, an [inverter](#) is used in solar power plants.
- For a large-scaled grid-tied power plant, the inverter is connected with special protective devices. And a [transformer](#) is also connected with the inverter to assures the output voltage and frequency as per the standard supply.

Energy storage devices

The batteries are used to store electrical energy generated by the solar power plants. The storage components are the most important component in a power plant to meet the demand and variation of the load. This component is used especially when the sunshine is not available for few days.

The capacity of a battery is that how much amount of electrical power it can store. The capacity of batteries is measured in Ampere-hours (AH) rating.

For example, a battery having 100 AH battery can supply 1 Amp current for 100 hours or 100 Amp current for 1 hour.

For a long life of a battery, never fully discharge a battery. And in case, if a battery is fully discharged, never keep fully discharged battery for a long time.

The capacity of a battery is affected by the temperature. There is a reduction of 0.6% of capacity for every degree Celsius rise in temperature more than 25° C.

There are two types of batteries used in the solar power plant;

Lead-Acid battery

Nickel-Cadmium battery

Charge Controller

A charge controller is used to control the charging and discharging of the battery. The charge controller is used to avoid the overcharging of the battery. The overcharging of a battery may lead to corrosion and reduce plate growth. And in the worst condition, it may damage the electrolyte of the battery.

Sometimes, the charge controller is termed a solar battery charger. There are many technologies used to make a charge controller. For example, the most popular technique is the MPPT charge controller that is known as "Maximum Power Point Tracking". This algorithm is used to optimize the production of PV cells.

TYPICAL AC POWER SUPPLY SCHEME

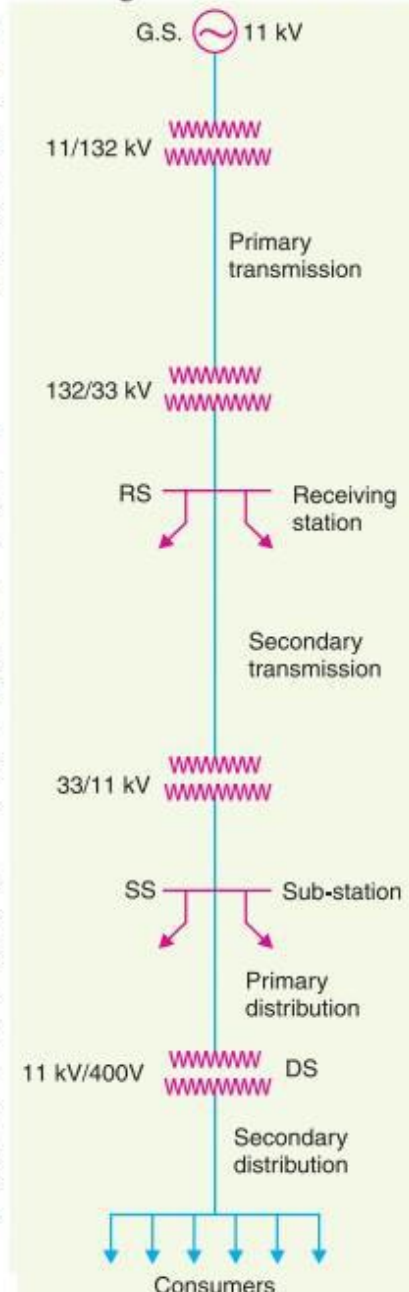
An electric supply system consists of three principal components *viz.*, the power station, the transmission lines and the distribution system. Electric power is produced at the power stations which are located at favourable places, generally quite away from the consumers. It is then transmitted over 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.

Typical a.c. Power Supply Scheme

The large network of conductors between the power station and the consumers can be broadly divided into two parts *viz.*, transmission system and distribution system. Each part can be further sub-divided into two—primary transmission and secondary transmission and primary distribution and secondary distribution. Fig. 7.1. shows the layout of a typical a.c. power supply scheme by a single line diagram. It may be noted that it is not necessary that all power schemes include all the stages shown in the figure. For example, in a certain power scheme, there may be no secondary transmission and in another case, the scheme may be so small that there is only distribution and no transmission.

(i) **Generating station :** In Fig 7.1, G.S. represents the generating station where electric power is produced by 3-phase alternators operating in parallel. The usual 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 transformers. 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 the highest possible voltage for transmission of electric power to save conductor material and have other advantages.



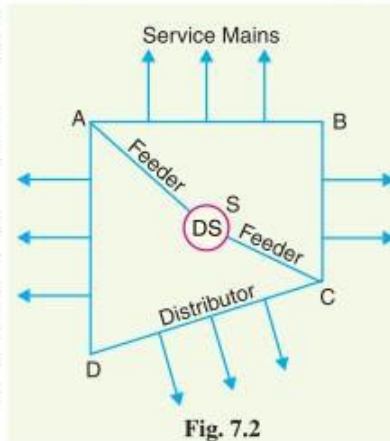
well as the cost of switchgear and transformer equipment is increased. Therefore, the choice of proper transmission voltage is essentially a question of economics. Generally the primary transmission is carried at 66 kV, 132 kV, 220 kV or 400 kV.

(ii) Primary transmission. The electric power at 132 kV is transmitted by 3-phase, 3-wire overhead system to the outskirts of the city. This forms the primary transmission.

(iii) 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 reduced to 33kV by step-down transformers. From this station, electric power is transmitted at 33kV by 3-phase, 3-wire overhead system to various sub-stations (SS) located at the strategic points in the city. This forms the secondary transmission.

(iv) 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 consumers (having demand more than 50 kW) are generally supplied power at 11 kV for further handling with their own sub-stations.

(v) 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.



Elements of a Transmission Line

For reasons associated with economy, transmission of electric power is done at high voltage by 3-phase, 3-wire overhead system. The principal elements of a high-voltage transmission line are :

- (i) Conductors**, usually three for a single-circuit line and six for a double-circuit line. The usual material is aluminium reinforced with steel.
- (ii) Step-up and step-down transformers**, at the sending and receiving ends respectively. The use of transformers permits power to be transmitted at high efficiency.
- (iii) Line insulators**, which mechanically support the line conductors and isolate them electrically from the ground.
- (iv) Support**, which are generally steel towers and provide support to the conductors.
- (v) Protective devices**, such as ground wires, lightning arrestors, circuit breakers, relays etc. They ensure the satisfactory service of the transmission line.
- (vi) Voltage regulating devices**, which maintain the voltage at the receiving end within permissible limits.

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 the following properties :

- (i) high electrical conductivity.
- (ii) high tensile strength in order to withstand mechanical stresses.
- (iii) low cost so that it can be used for long distances.
- (iv) 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 overhead 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.

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 :

- (i) High mechanical strength to withstand the weight of conductors and wind loads etc.
- (ii) Light in weight without the loss of mechanical strength.
- (iii) Cheap in cost and economical to maintain.
- (iv) Longer life.
- (v) Easy accessibility of conductors for maintenance.

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.

Insulators

The overhead line conductors should be supported on the poles or towers in such a way that currents from conductors do not flow to earth through supports *i.e.*, line conductors must be properly 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 :

- (i) High mechanical strength in order to withstand conductor load, wind load etc.
- (ii) High electrical resistance of insulator material in order to avoid leakage currents to earth.
- (iii) High relative permittivity of insulator material in order that dielectric strength is high.
- (iv) The insulator material should be non-porous, free from impurities and cracks otherwise the permittivity will be lowered.
- (v) High ratio of puncture strength to flashover.

The most commonly used material for insulators of overhead line is *porcelain* but glass, steatite and special composition materials are also used to a limited extent. Porcelain is produced by firing at a high temperature a mixture of kaolin, feldspar and quartz. It is stronger mechanically than glass, gives less trouble from leakage and is less effected by changes of temperature.

Location of Voltage Control Equipment

In a modern power system, there are several elements between the generating station and the consumers. The voltage control equipment is used at more than one point in the system for two reasons. Firstly, the power network is very extensive and there is a considerable voltage drop in transmission and distribution systems. Secondly, the various circuits of the power system have dissimilar load characteristics. For these reasons, it is necessary to provide individual means of voltage control for each circuit or group of circuits. In practice, voltage control equipment is used at :

- (i) generating stations
- (ii) transformer stations
- (iii) the feeders if the drop exceeds the permissible limits

Methods of Voltage Control

There are several methods of voltage control. In each method, the system voltage is changed in accordance with the load to obtain a fairly constant voltage at the consumer's end of the system. The following are the methods of voltage control in an *a.c.* power system:

- (i) By excitation control
- (ii) By using tap changing transformers
- (iii) Auto-transformer tap changing
- (iv) Booster transformers
- (v) Induction regulators
- (vi) By synchronous condenser

Power Transformers. A power transformer is used in a sub-station to step-up or step-down the voltage. all the subsequent sub-stations use step-down transformers to gradually reduce the voltage of electric supply and finally deliver it at utilisation voltage. The modern practice is to use 3-phase transformers in sub-stations

Protective Devices:

Switches. A switch is a device which is used to open or close an electrical circuit in a convenient way. It can be used under full-load or no-load conditions *but* it cannot interrupt the fault currents. When the contacts of a switch are opened, an **arc* is produced in the air between the contacts. This is particularly true for circuits of high voltage and large current capacity. The switches may be classified into (i) air switches (ii) oil switches. The contacts of the former are opened in air and that of the latter are opened in oil.

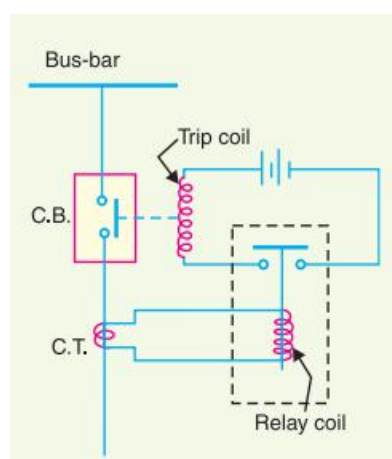
Fuses. A fuse is a short piece of wire or thin strip which melts when excessive current flows through it for sufficient time. It is inserted in series with the circuit to be protected. Under normal operating conditions, the fuse element is at a temperature below its melting point. Therefore, it carries the normal load current without overheating.

Instrument transformers. The lines in sub-stations operate at high voltages and carry current of thousands of amperes. The measuring instruments and protective devices are designed for low voltages (generally 110 V) and currents (about 5 A). Therefore, they will not work satisfactorily if mounted directly on the power lines. This difficulty is overcome by installing *instrument transformers* on the power lines. The function of these instrument transformers is to transfer voltages or currents in the power lines to values which are convenient for the operation of measuring instruments and relays. There are two types of instrument transformers *viz.*

- (i) Current transformer (C.T.)
- (ii) Potential transformer (P.T.)

Circuit breakers. A circuit breaker is an equipment which can open or close a circuit under all conditions *viz.* no load, full load and fault conditions. It is so designed that it can be operated manually (or by remote control) under normal conditions and automatically under fault conditions. For the latter operation, a relay circuit is used with a circuit breaker.

Relays. A relay is a device which detects the fault and supplies information to the breaker for circuit interruption.



Under normal load conditions, the e.m.f. of the secondary winding of C.T. is small and the current flowing in the relay operating coil is insufficient to close the relay contacts. This keeps the trip coil of the circuit breaker unenergised. Consequently, the contacts of the circuit breaker remain closed and it carries the normal load current. When a fault occurs, a large current flows through the primary of C.T. This increases the secondary e.m.f. and hence the current through the relay operating coil. The relay contacts are closed and the trip coil of the circuit breaker is energised to open the contacts of the circuit breaker.

TYPES OF DISTRIBUTION SYSTEMS: PRIMARY AND SECONDARY DISTRIBUTION SYSTEMS

Now-a-days electrical energy is generated, transmitted and distributed in the form of alternating current. One important reason for the widespread use of alternating current in preference to direct current is the fact that alternating voltage can be conveniently changed in magnitude by means of a transformer. Transformer has made it possible to transmit a.c. power at high voltage and utilise it at a safe potential. High transmission and distribution voltages have greatly reduced the current in the conductors and the resulting line losses.

There is no definite line between transmission and distribution according to voltage or bulk capacity. However, in general, the a.c. distribution system is the electrical system between the step-down substation fed by the transmission system and the consumers' meters. The a.c. distribution system is classified into (i) primary distribution system and (ii) secondary distribution system.

- (i) **Primary distribution system.** It is that part of a.c. distribution system which operates at voltages somewhat higher than general utilisation and handles large blocks of electrical energy than the average low-voltage consumer uses. The voltage used for primary distribu-

tion depends upon the amount of power to be conveyed and the distance of the substation required to be fed. The most commonly used primary distribution voltages are 11 kV, 6.6 kV and 3.3 kV. Due to economic considerations, primary distribution is carried out by 3-phase, 3-wire system.

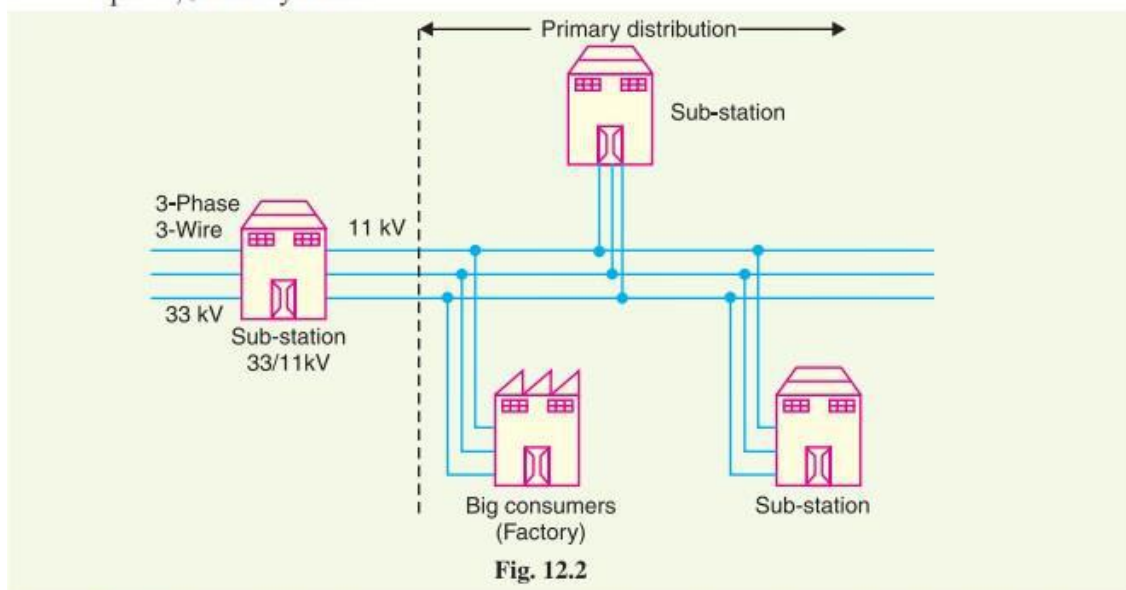


Fig. 12.2 shows a typical primary distribution system. Electric power from the generating station is transmitted at high voltage to the substation located in or near the city. At this substation, voltage is stepped down to 11 kV with the help of step-down transformer. Power is supplied to various substations for distribution or to big consumers at this voltage. This forms the high voltage distribution or primary distribution.

- (ii) **Secondary distribution system.** It is that part of a.c. distribution system which includes the range of voltages at which the ultimate consumer utilises the electrical energy delivered to him. The secondary distribution employs 400/230 V, 3-phase, 4-wire system.

Fig. 12.3 shows a typical secondary distribution system. The primary distribution circuit delivers power to various substations, called distribution substations. The substations are situated near the consumers' localities and contain step-down transformers. At each distribution substation, the voltage is stepped down to 400 V and power is delivered by 3-phase, 4-wire a.c. system. The voltage between any two phases is 400 V and between any phase and neutral is 230 V. The single phase domestic loads are connected between any one phase and the neutral, whereas 3-phase 400 V motor loads are connected across 3-phase lines directly.



Power transformer

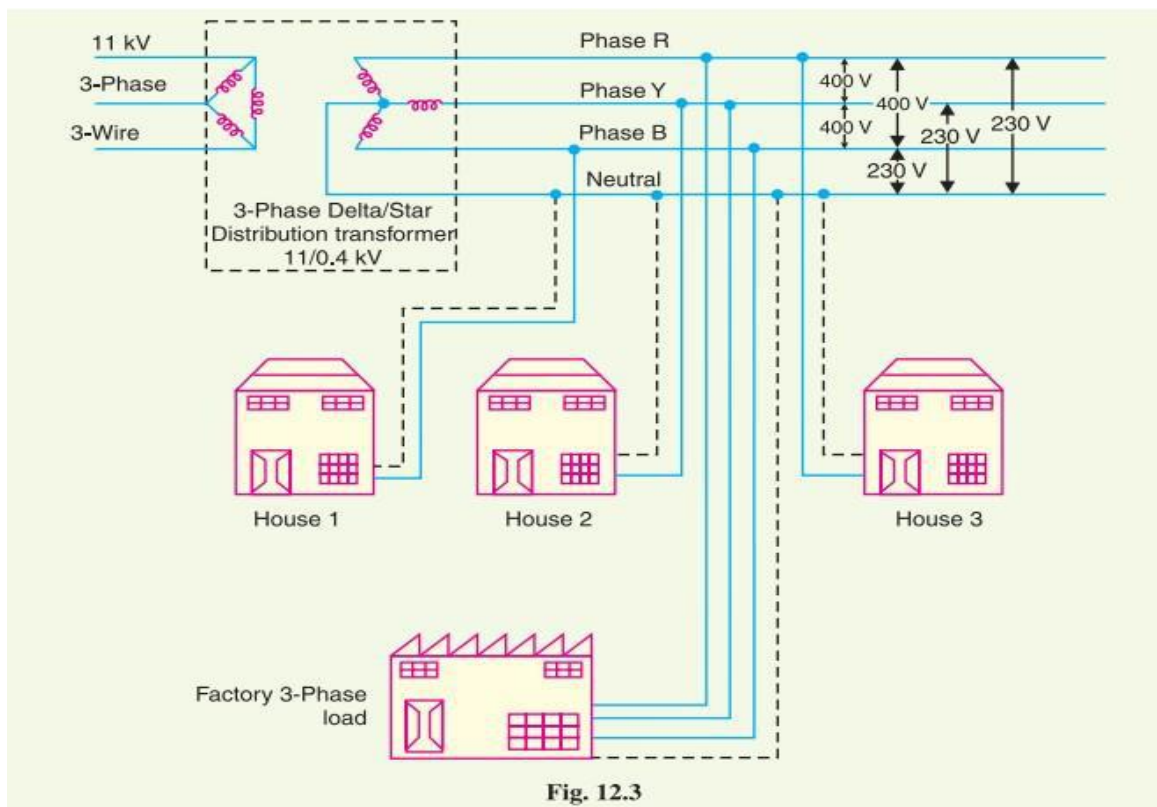


Fig. 12.3

Fig. Secondary Distribution Systems

***** THE END*****

UNIT-1

Analog Electronics

①

⇒ Overview of Semiconductors:

* Electronics deals with the flow of electrons through vacuum, gas or semiconductors.

* Electronic devices like diodes, transistors, resistors, capacitors etc form circuits of electronic gadgets.

* An electronic device consists of integrated circuits which have several diodes, transistors, resistors, capacitors etc mounted on a single chip.

* Electronic components like diodes & transistors are made of semiconductor materials.

* All materials are classified into 3 types

1. Conductors

2. Semiconductors

3. Insulators

⇒ Conductors: These materials have large numbers of free electrons in their atomic structure which allows flow of current. Materials which allow electric current to flow.

Ex: Gold, Silver, Copper, Aluminium

⇒ Semiconductors: A semiconductor has an electrical conductivity between a conductor and an insulator.

Its ability to conduct electricity increases with increase in temperature or by applying voltage.

Ex: Silicon, Germanium

⇒ Insulators: These materials do not have free electrons, hence current does not flow through them.

Ex: Rubber, Glass, Wood

Materials which do not allow electric current to flow are called insulators.

⇒ Atomic Theory:

- * Atom of any material consists of a central nucleus around which electrons are orbiting in different orbits.
- * Electrons are held in orbits due to electrostatic force between electrons and the nucleus.
- * Electrons are negatively charged whereas protons in the nucleus are positively charged.
- * As there are equal number of protons (+ve charge) & electrons (-ve charge), an atom is electrically neutral.
- * When an atom loses an electron, it will lose some -ve charge & will become a +vely charged ion.
- * When an atom gains an electron it will gain some -ve charge & will become a -vely charged ion.

⇒ Atomic Structure of Silicon & Germanium Atom:

* The electrons present in the outermost shell are called valence electrons.

* Silicon has 14 electrons orbiting in 3 orbits & the electrons are distributed as 2, 8 & 4.

* Germanium has 32 electrons orbiting in 4 orbits & the electrons are distributed as 2, 8, 18, 4.

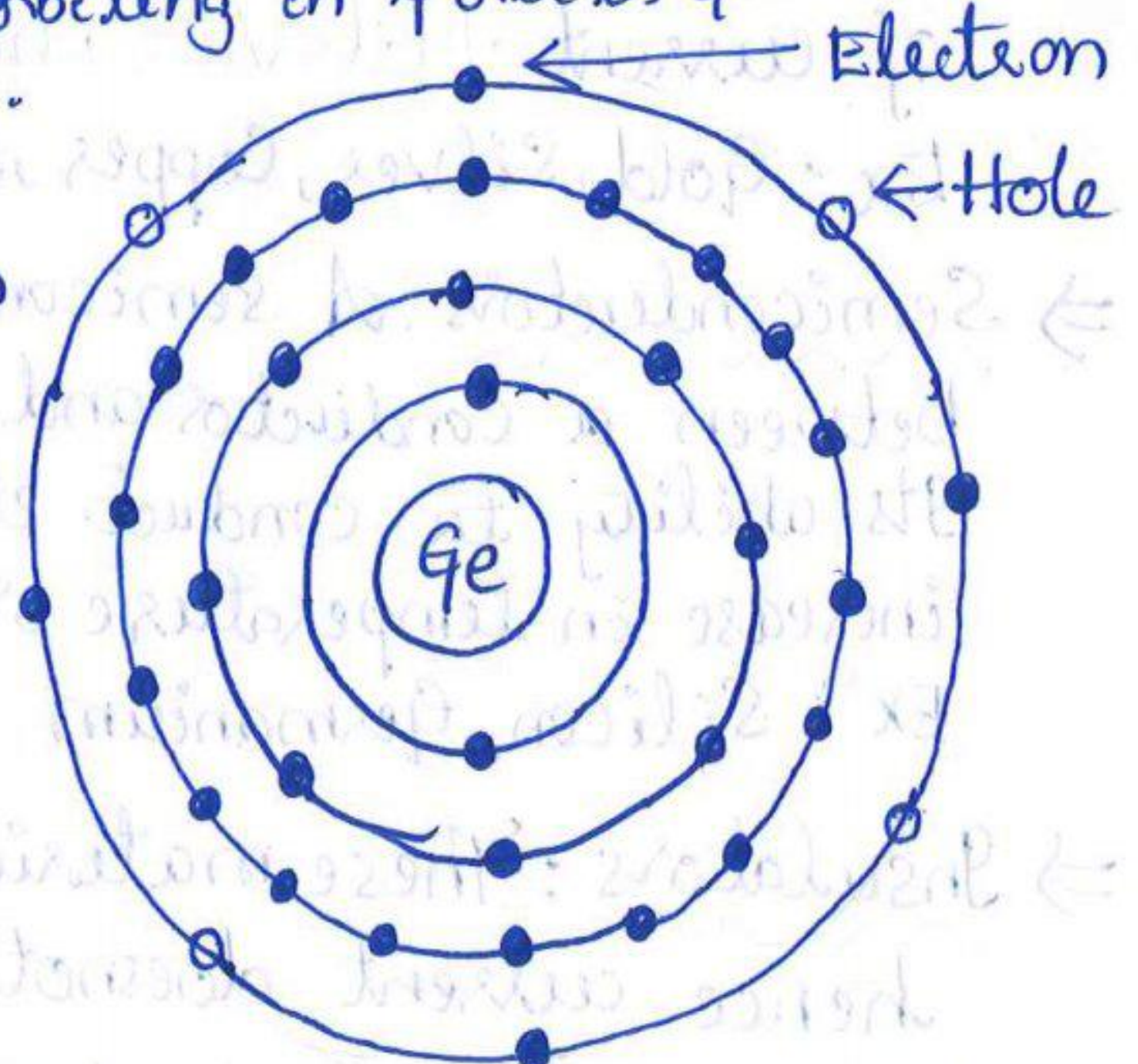
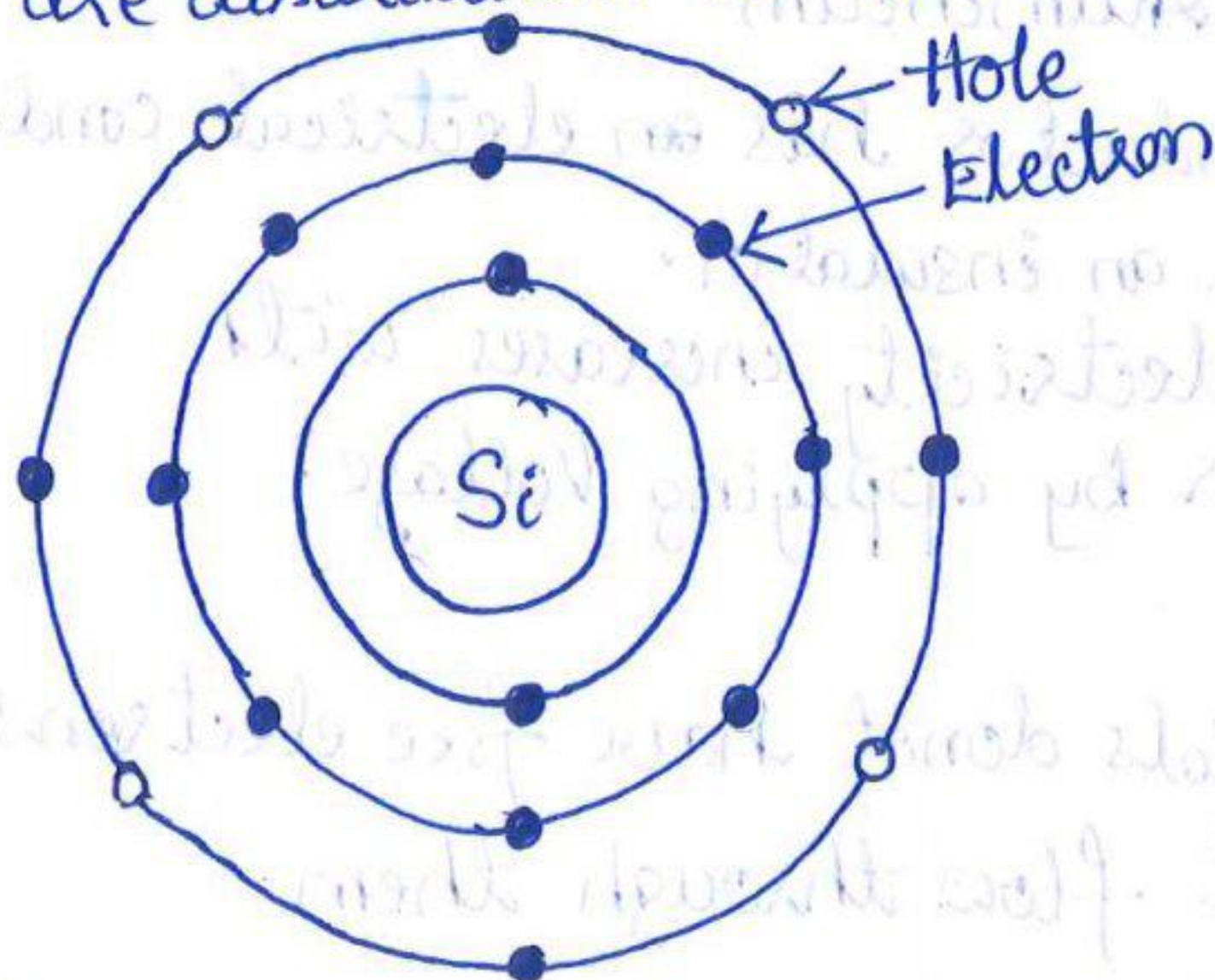


Fig: Atomic Structure of Silicon & Germanium Atom

- ✗ Both Silicon & Germanium have 4 valence electrons and 4 holes in the outermost orbit.
- ✗ Electrons in the outermost orbit are loosely bound to the nucleus.
- ✗ A small amount of energy is required to take out an electron from the outermost orbit.
- ✗ When an electron leaves its orbit, it becomes free electron.

⇒ Covalent Bond:

- ✗ A semiconductor atom having four valence electrons & four holes require four more electrons so as to make the outermost orbit completely filled.
- ✗ Covalent bond is the chemical bond that involves the sharing of electron pairs between atoms.
- ✗ Sharing of electrons of the neighbouring atoms to satisfy the need to have 8 electrons in the valence orbit of an atom is called covalent bonding.
- ✗ Due to covalent bonding, the outermost orbit is full.
- ✗ At absolute zero temperature, there will be no free electrons in the crystal.
- ✗ Though the electrons are bound to their atoms due to covalent bonding, increase in the temperature breaks some of the covalent bonding & makes some electrons free.

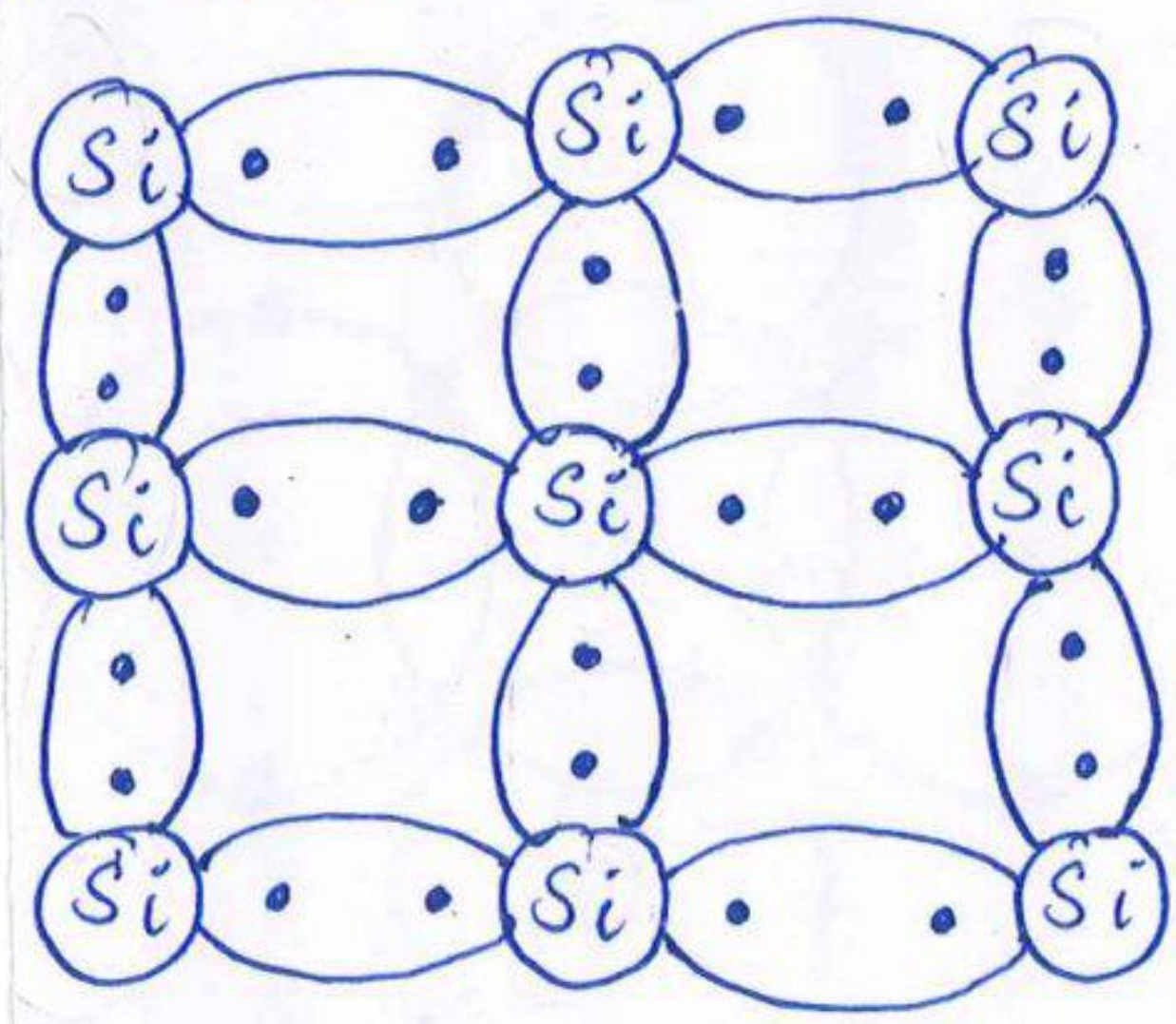


Fig: Covalent Bonding of Silicon Atoms

⇒ Types of Semiconductors:

Semiconductors are divided into 2 types.

1. Intrinsic Semiconductors
2. Extrinsic Semiconductors

* Intrinsic Semiconductor:

An intrinsic Semiconductor is a pure ^{form of} semiconductor.

* Conductivity of the intrinsic semiconductor is zero at room temperature.

* Semiconductors in which electrons are bound to their atoms and are not free to conduct electric current are called intrinsic semiconductors.

* When the temperature increases, the heat energy enables the valence electrons to get sufficient energy to break the bond and become free electrons.

* When an electron becomes free, it leaves a vacant space called a hole.

* For every free electron, there will be a hole produced, which is called electron-hole pair.

* Due to rise in temperature, electron-hole pairs are formed.

* The free electrons get attracted & fall into a hole created by another electron.

⇒ Doping:

* The conductivity of semiconductor materials can be increased by adding another material having 3 or 5 valence electrons.

* The process of adding impurities to the pure semiconductor is called doping.

* The process of adding either a pentavalent or trivalent element to a pure semiconductor is called doping.

⇒ Extrinsic Semiconductors:

A semiconductor to which an impurity is added to make it conductive is called an extrinsic semiconductor.

* The doped semiconductor is called extrinsic semiconductor.

* The doping material that is added to a pure semiconductor is called impurity material.

⇒ Types of Extrinsic Semiconductors:

Depending upon the type of impurity that is added, extrinsic semiconductors are divided into 2 types:

1. P-Type Semiconductors

2. N-Type Semiconductors

⇒ P-Type Semiconductor:

* When a trivalent impurity is added to the intrinsic semiconductor, a P-Type semiconductor is formed.

* Example of trivalent impurity materials: Boron, Gallium, Indium, Aluminium

* When a trivalent impurity Boron is added to Silicon

* Covalent bonds are formed between Boron & Silicon Boron having 3 valence electrons & Silicon having 4 valence electrons.

* There will be shortage of one electron in the covalent bonds This is represented by an empty space called a hole.

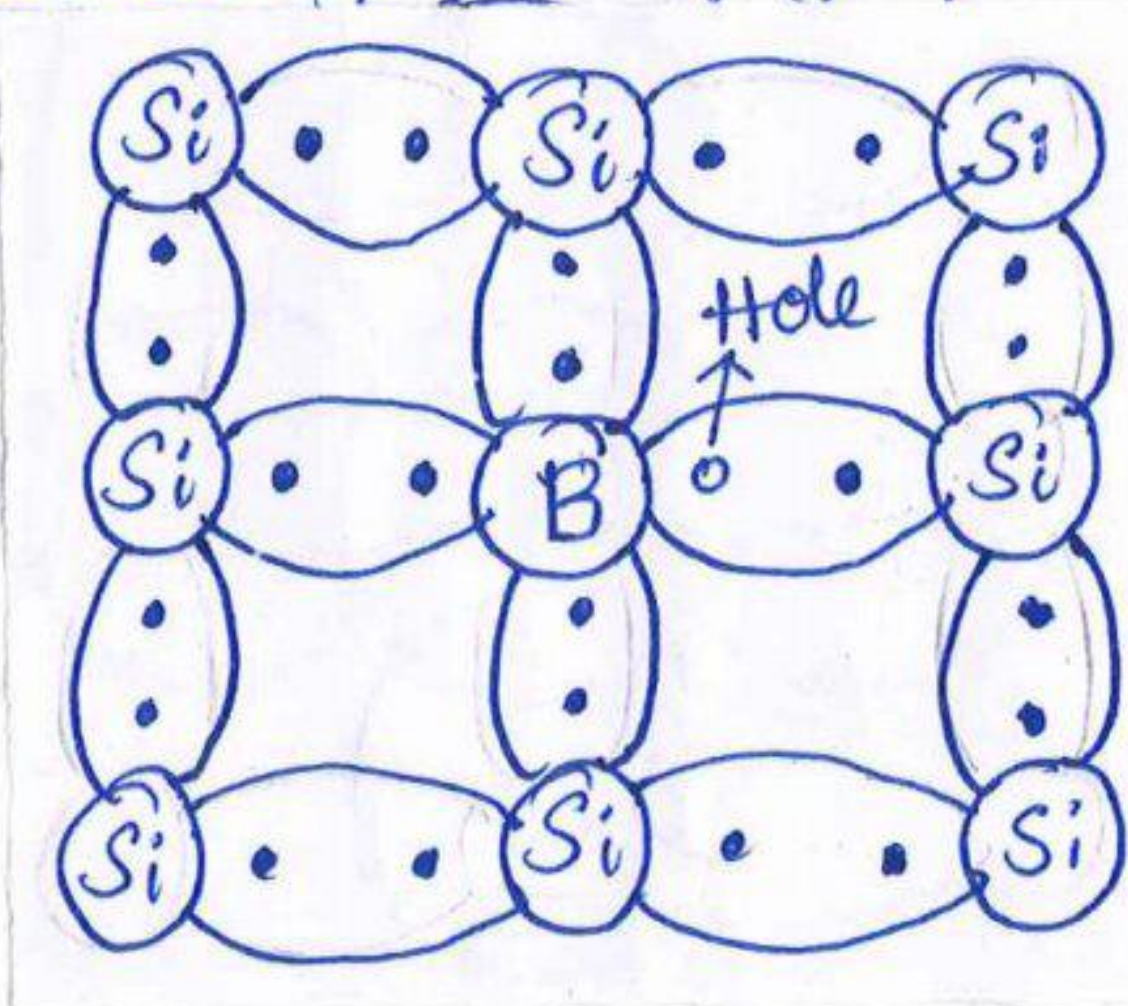
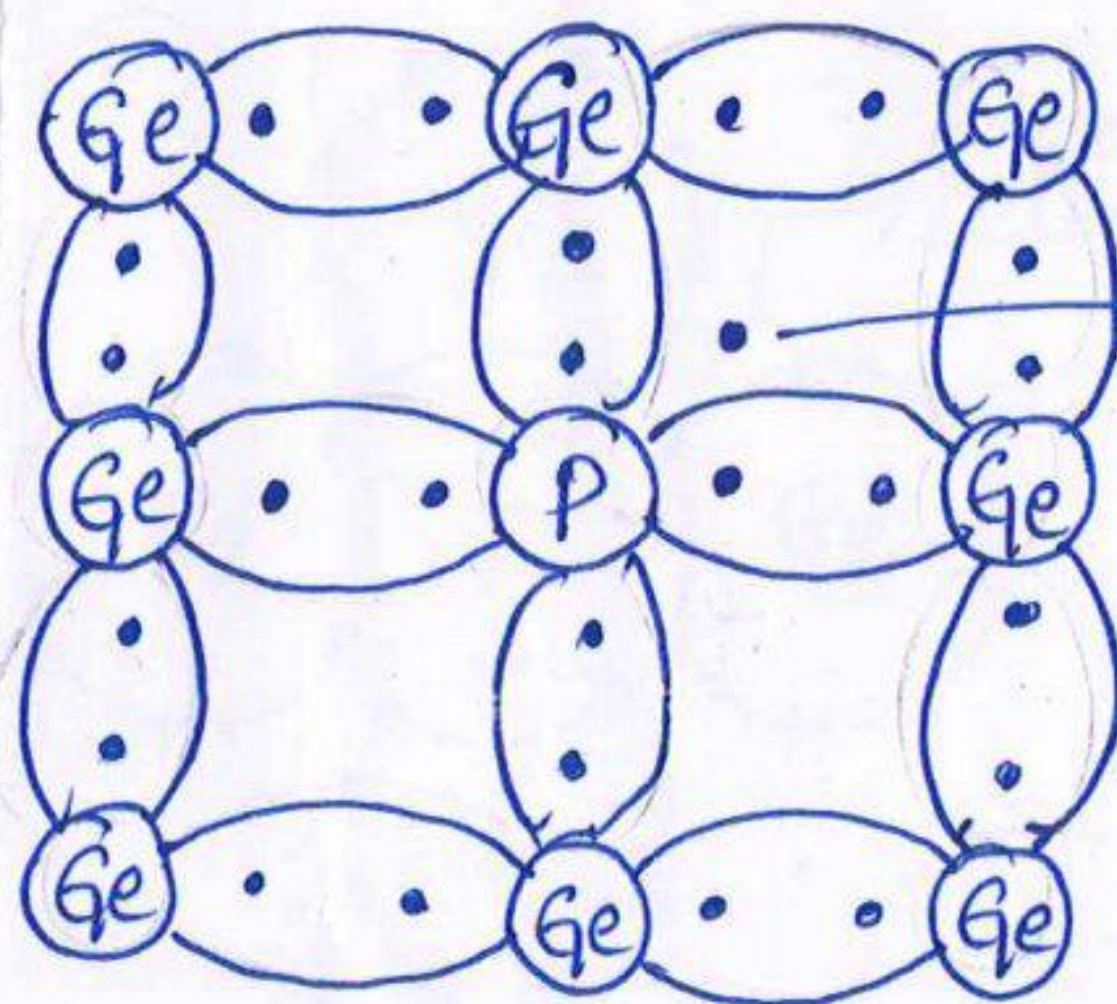


Fig: P-Type Semiconductor Material

- * Based on the total number of impurity atoms that are added, large number of holes are produced.
- * Hence a p-type material will have plenty of holes.
- * In a p-type semiconductor, holes are majority charge carriers and electrons are minority charge carriers.
- * As holes are treated as positively charged, the material is known as P-type material.
- * In a P-type Semiconductor, conductivity is because of positively charged holes.

⇒ N-Type Semiconductor:

- * When a pentavalent impurity is added to the intrinsic semiconductor, N-type semiconductor is formed.
- * Example of Pentavalent impurity materials: Phosphorous, Arsenic, Antimony
- * When a pentavalent impurity Phosphorous is added to Germanium
 - ↓
 - * Covalent Bonds are formed between Phosphorous & Germanium. Phosphorous has 5 valence electrons & Germanium has 4 valence electrons.
 - ↓
 - * Four out of five valence electrons will form covalent bonds with Germanium atoms having four valence electrons & one electron is left free.
 - ↓
 - * For each Phosphorous atom, there will be one free electron.



Free electron

Fig: N-Type Semiconductor

- * The free electrons which doesnot take part in covalent bonding are free to conduct electricity.
- * A N-type material will have large number of free electrons.
- * As electrons are treated as negatively charged carriers, the material is known as N-type material.
- * In a N-type semiconductor, conductivity is because of negatively charged electrons.
- * In a N-type Semiconductor, electrons are majority charge carriers & holes are minority charge carriers.

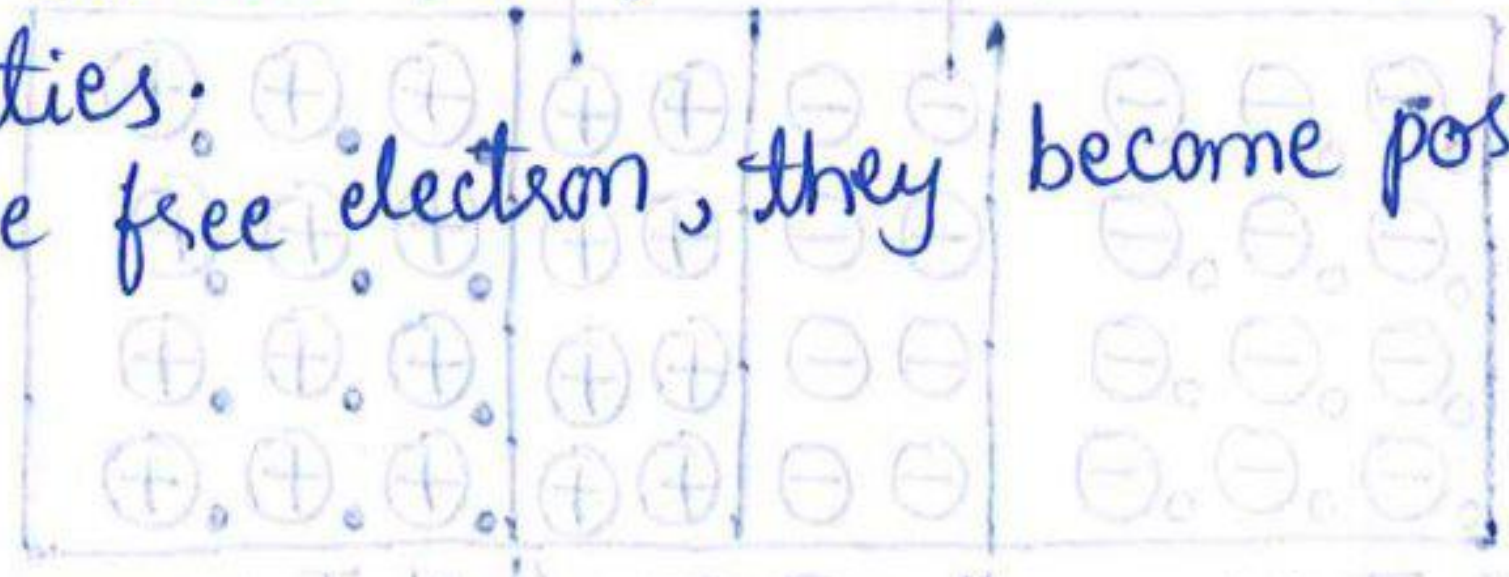
⇒ Acceptor Impurities & Acceptor Ions: carriers

- * A P-type semiconductor has holes as majority charge
- * As holes are ready to accept free electrons, the trivalent impurities that produce a p-type semiconductor are called acceptor impurities.
- * As the free electron occupies a hole, it creates -ve ions
- * The number of electrons with -ve charge become more than the number of protons with +ve charge.
- * Hence a P-type Semiconductor will have -ve acceptor ions.

⇒ Donor Impurities & Donor Ions: carriers

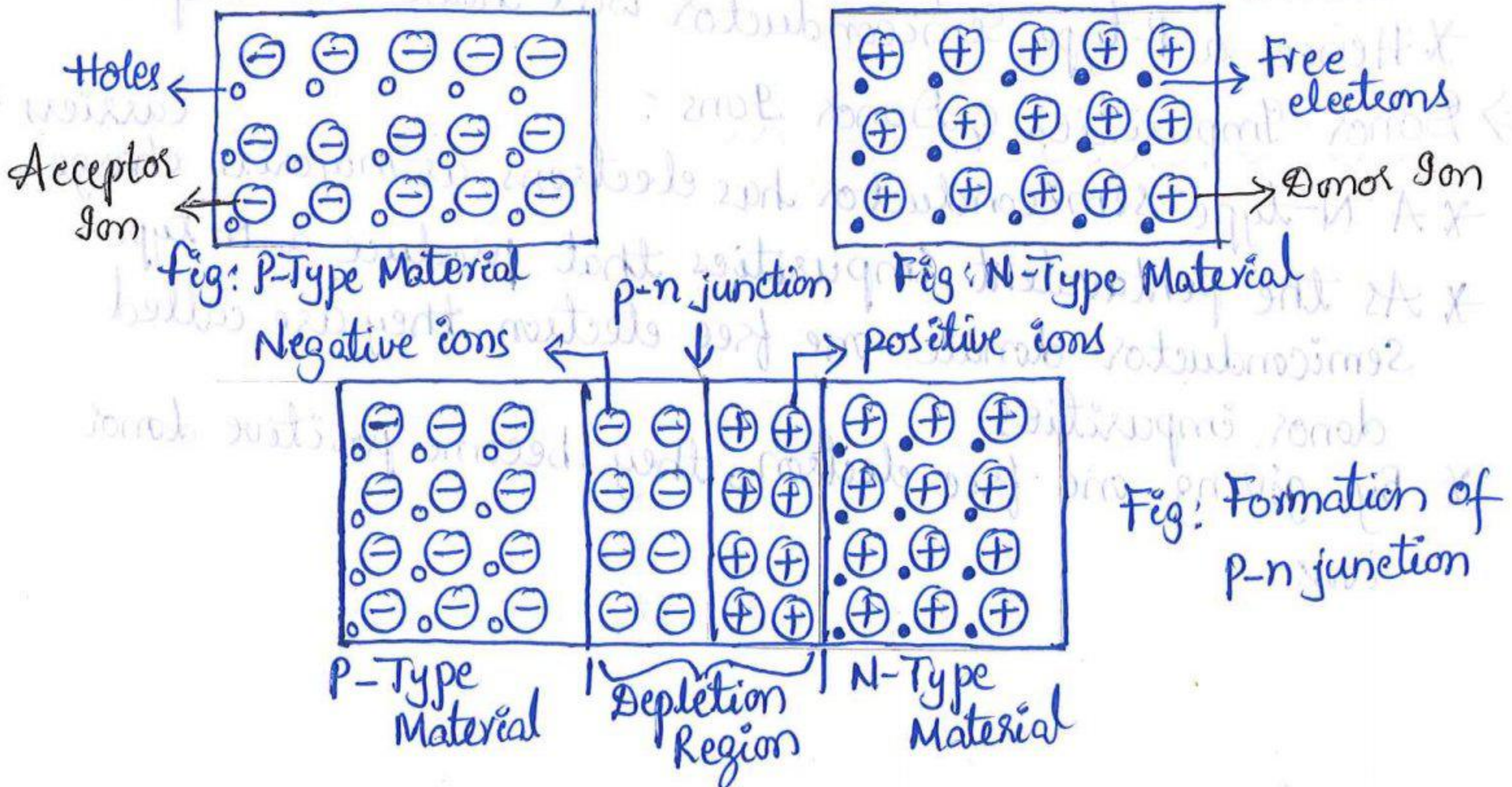
- * A N-type semiconductor has electrons as majority charge
- * As the pentavalent impurities that produce a N-type semiconductor donate one free electron, they are called donor impurities.

* By giving one free electron, they become positive donor ions.



⇒ P-N Junction:

- * When a P-type Semiconductor is joined with a N-type Semiconductor, the contact that is established between them is called PN junction.
- * The P-type material has high concentration of holes & the N-type material has high concentration of electrons.
- * At the pn junction, free electrons from the N-type material diffuse (move) into the p-side and the free electrons combine with the holes that are near the junction.
- * When an electron moves out of an atom, the atom becomes +vely charged immobile (unable to move) ion.
- * The free electron moving from n-side to p-side will leave positive immobile ions on the n-side of the junction.
- * The free electrons which cross the junction will occupy the holes in the p-type material making the atoms -vely charged immobile ions.
- (Atoms accepting -ve charge carriers become -vely charged ions)
- * On one side of the junction -ve ions are created & on the other side +ve ions are formed.



- * Negative ions created on the p-side will acquire negative voltage.
- * This -ve voltage on p-side will stop further diffusion of electrons from the n-side.
- * Positive ions created on the N-side will acquire positive voltage.
- * This +ve voltage on N-side will stop further diffusion of holes from the p-side.
- * Diffusion: It is defined as the movement of charge carriers from a high-concentration area to a low concentration area.
- * When a pn junction is formed, initially diffusion takes place & barrier voltage is created across the junction.
- * Barrier voltage stops further diffusion of charge carriers.
- * Barrier voltage depends upon amount of doping, charge carriers & junction temperature.
- * Depletion Region is formed at the junction.
- * Depletion region is the region which is depleted of charge carriers i.e., no mobile charge carriers are present in this region.
- * If the p-type and n-type materials are equally doped, then the depletion layer is equally divided on both sides of the pn junction.

⇒ P-N Junction Diode:

Diode = Di + Electrode [It has 2 electrodes]

- * A p-n junction has 2 terminals called electrodes.
- * The P-region acts as Anode & N-region acts as Cathode.

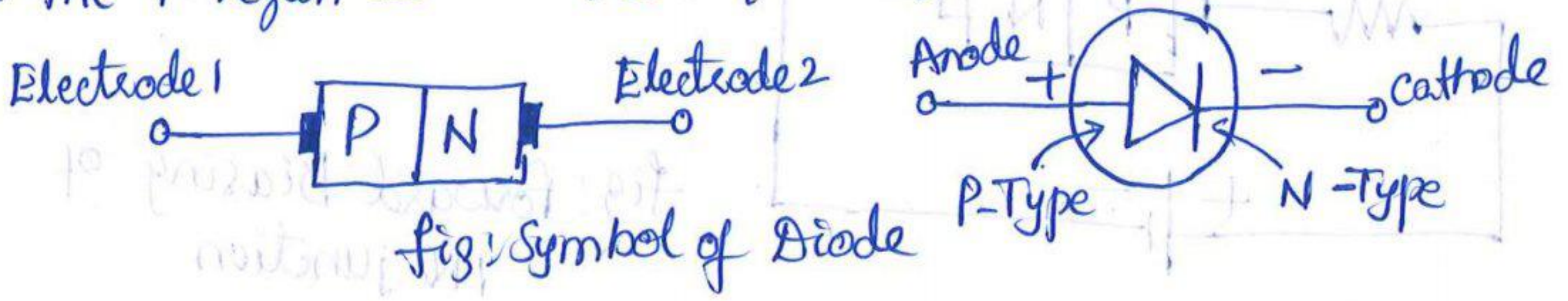


Fig: Symbol of Diode

* The arrow head in pn junction diode indicates the direction of current flow.

⇒ Biasing of P-N Junction Diode:

* Applying external voltage to any electronic device is called biasing.

* Definition: Applying some external voltage across two sides of the p-n junction is called biasing of p-n junction.

Types of Biasing: Depending upon the polarity of the externally applied voltage, biasing is classified as

1. Forward Biasing
2. Reverse Biasing

⇒ Forward Biased P-n junction:

* When the positive terminal of the battery is connected to the p-side and the negative terminal of the battery is connected to the n-side, the p-n junction is said to be a forward biased p-n junction.

* The holes on the p-side are positively charged & the electrons on the n-side are negatively charged.

* In forward biased condition, the positive terminal of the battery will repel the holes from the terminal.

* The electrons on n-side get repelled by the negative terminal of the battery.

* As a result, the width of the depletion region will be reduced. Potential barrier gets reduced.

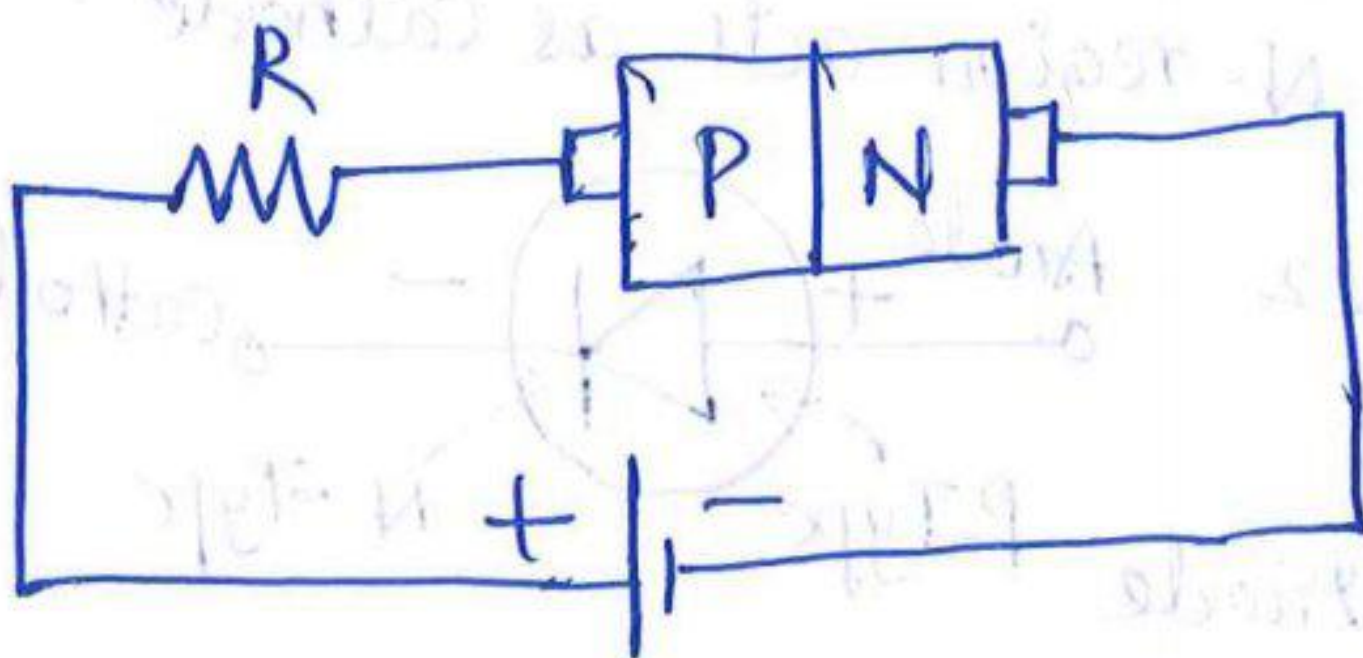


Fig: Forward Biasing of PN junction

- * If the applied voltage is still increased, the depletion region & barrier potential will disappear.
- * Resistor is connected to limit the flow of current.
 - ↳ reduces the flow of current.
- * When the voltage is gradually increased, when it reaches 0.3V, barrier voltage is overcome & depletion layer disappears for Germanium diode. For silicon potential barrier gets overcome at 0.7V.
- * Electrons on N-side get attracted by the +ve terminal of the battery & holes on P-side get attracted by the -ve terminal of the battery.
- * Majority charge carriers start moving across the p-n junction causing forward current I_f to flow.
- * A forward biased p-n junction offers very low resistance to current flow.

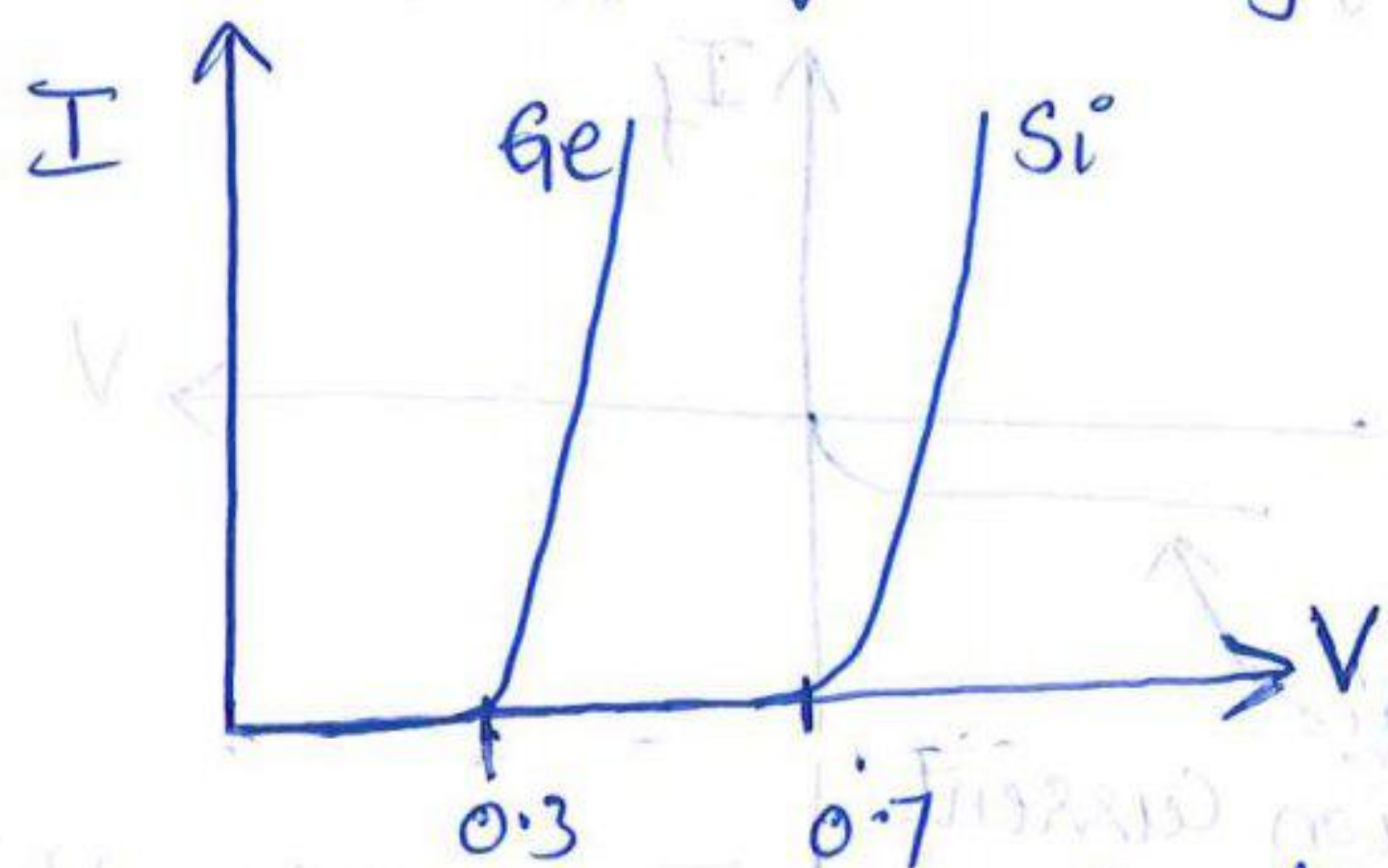
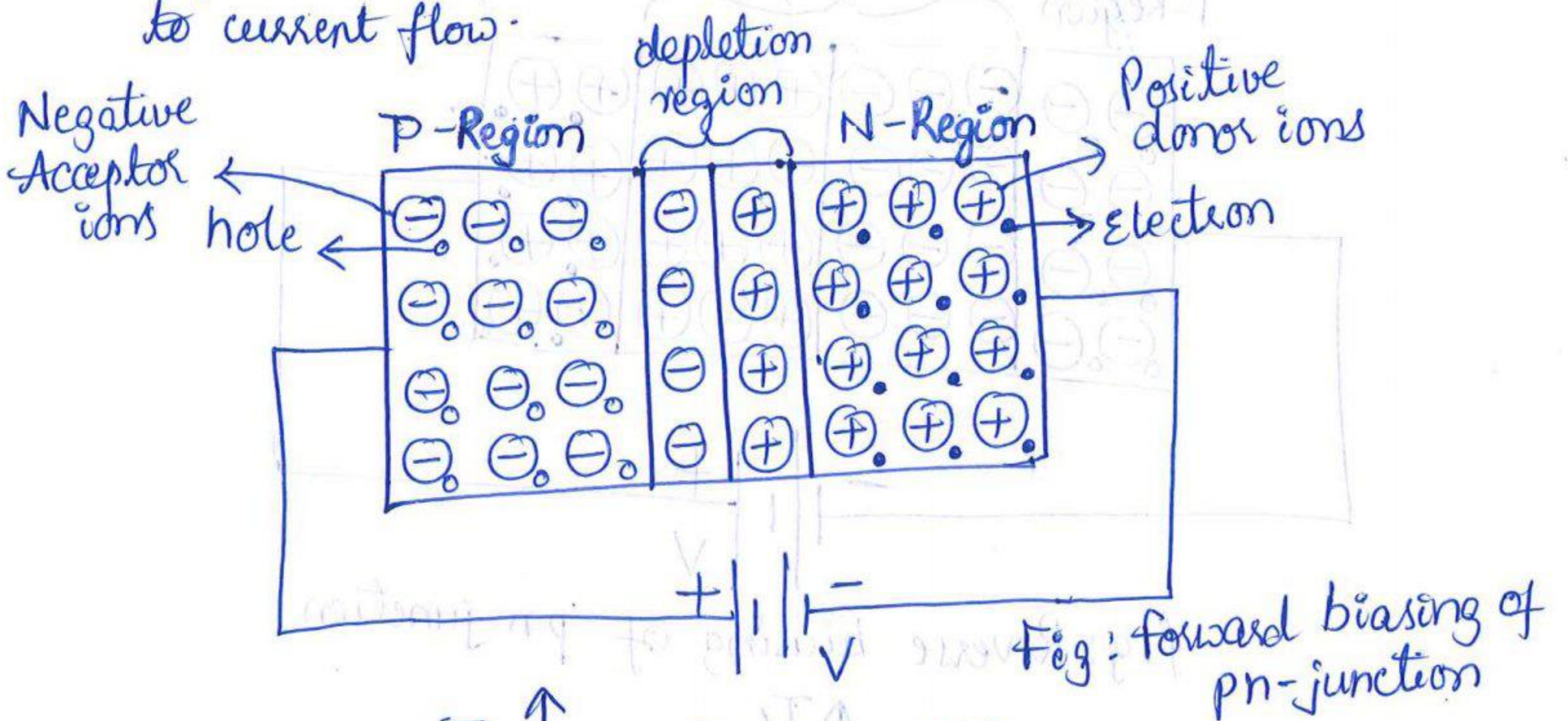


Fig: V-I characteristics of forward bias

⇒ Reverse Biased P-N Junction:

- ✖ When the negative terminal of the battery is connected to the P-side & the positive terminal of the battery is connected to the N-side, the P-N junction is called reverse biased P-N junction.
- ✖ Electrons on the n-side are attracted towards the +ve terminal of the battery & holes on the p-side gets attracted towards -ve terminal of the battery.
- ✖ As the applied voltage increases, depletion layer gets widened & the majority carriers don't cross the barrier.
- ✖ Due to minority charge carriers a small amount of current flows in the circuit which is called reverse saturation current.
- ✖ A reverse biased p-n junction offers very high resistance to current flow.

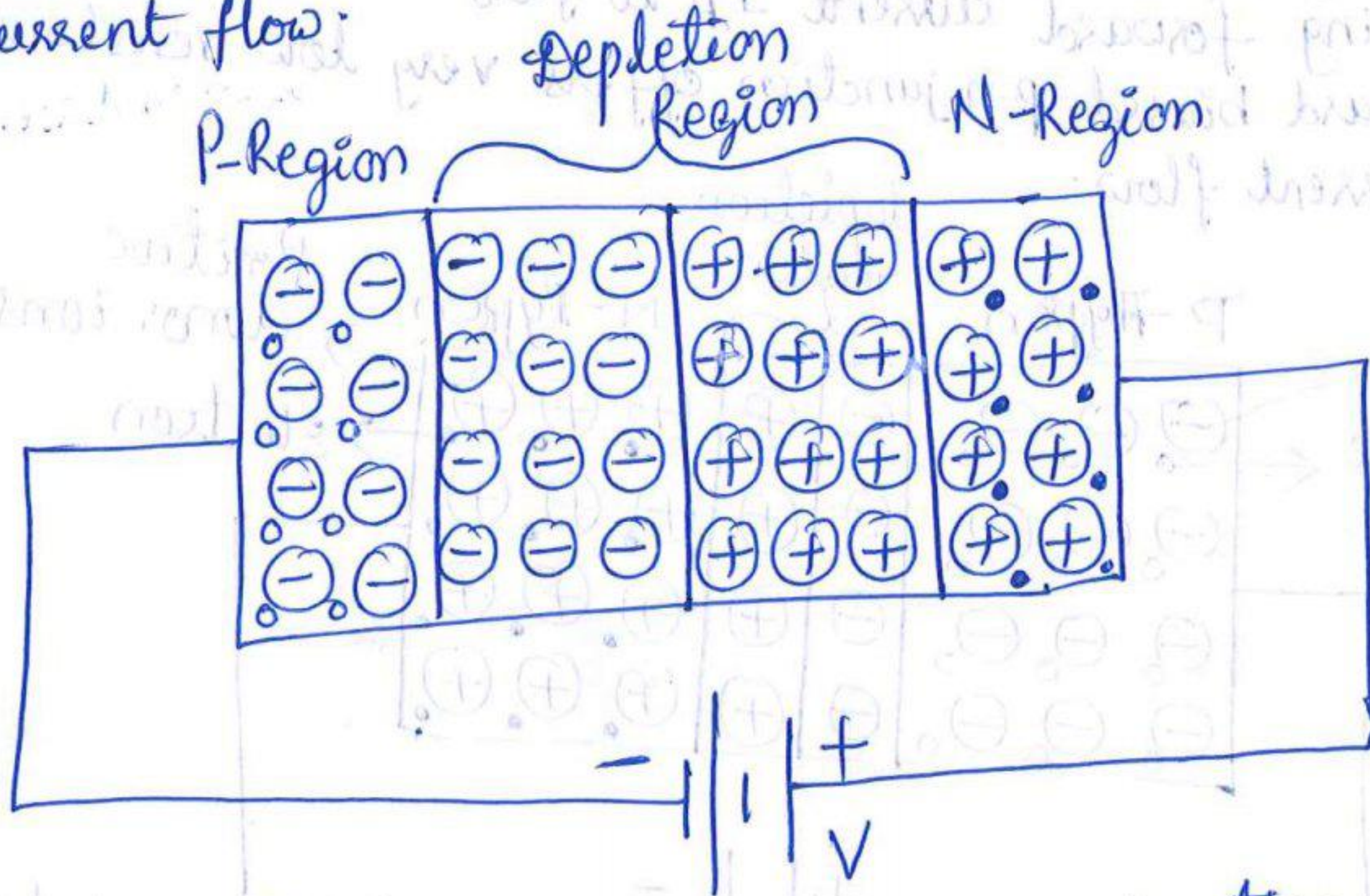


fig: Reverse biasing of p-n junction

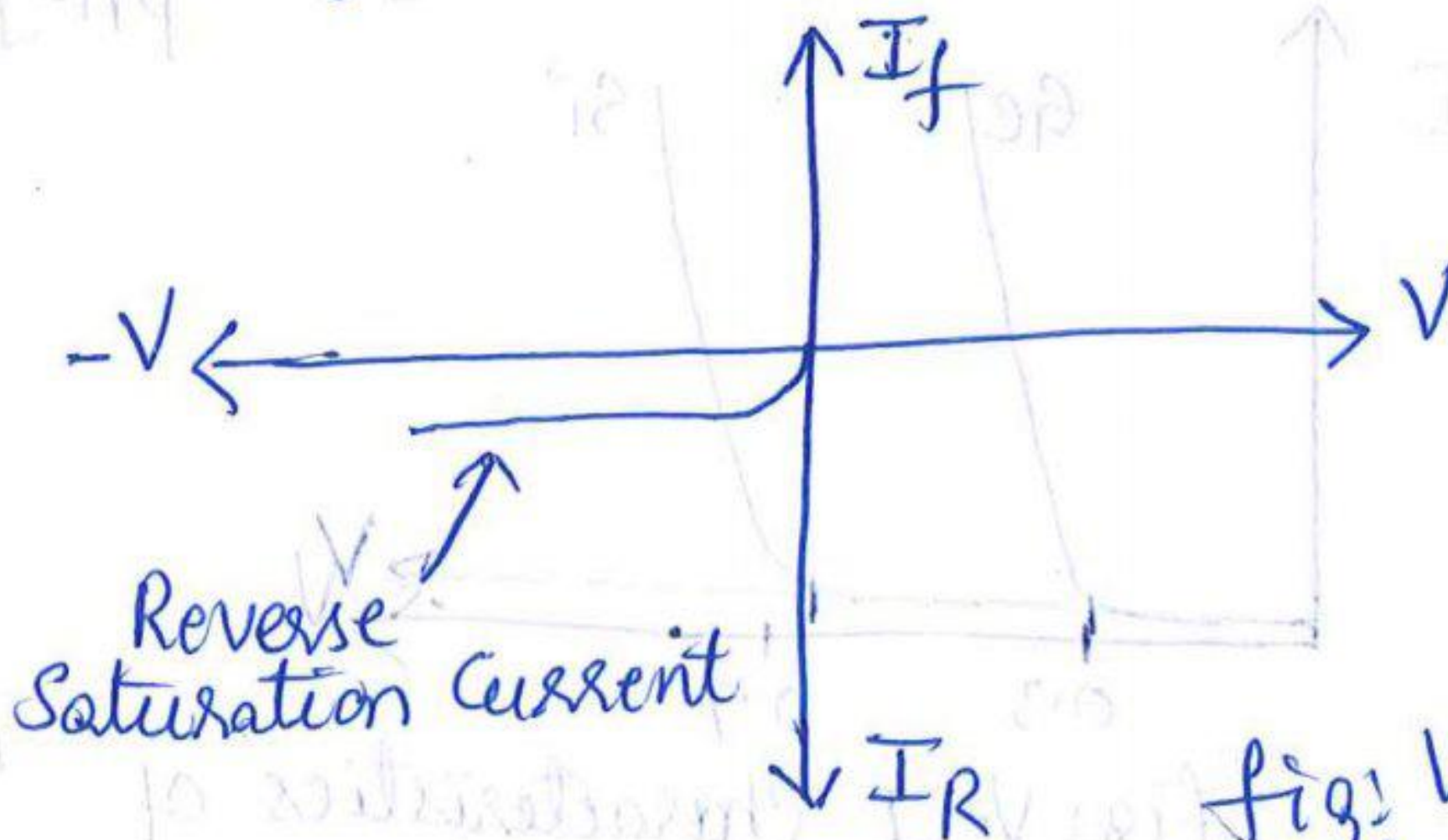
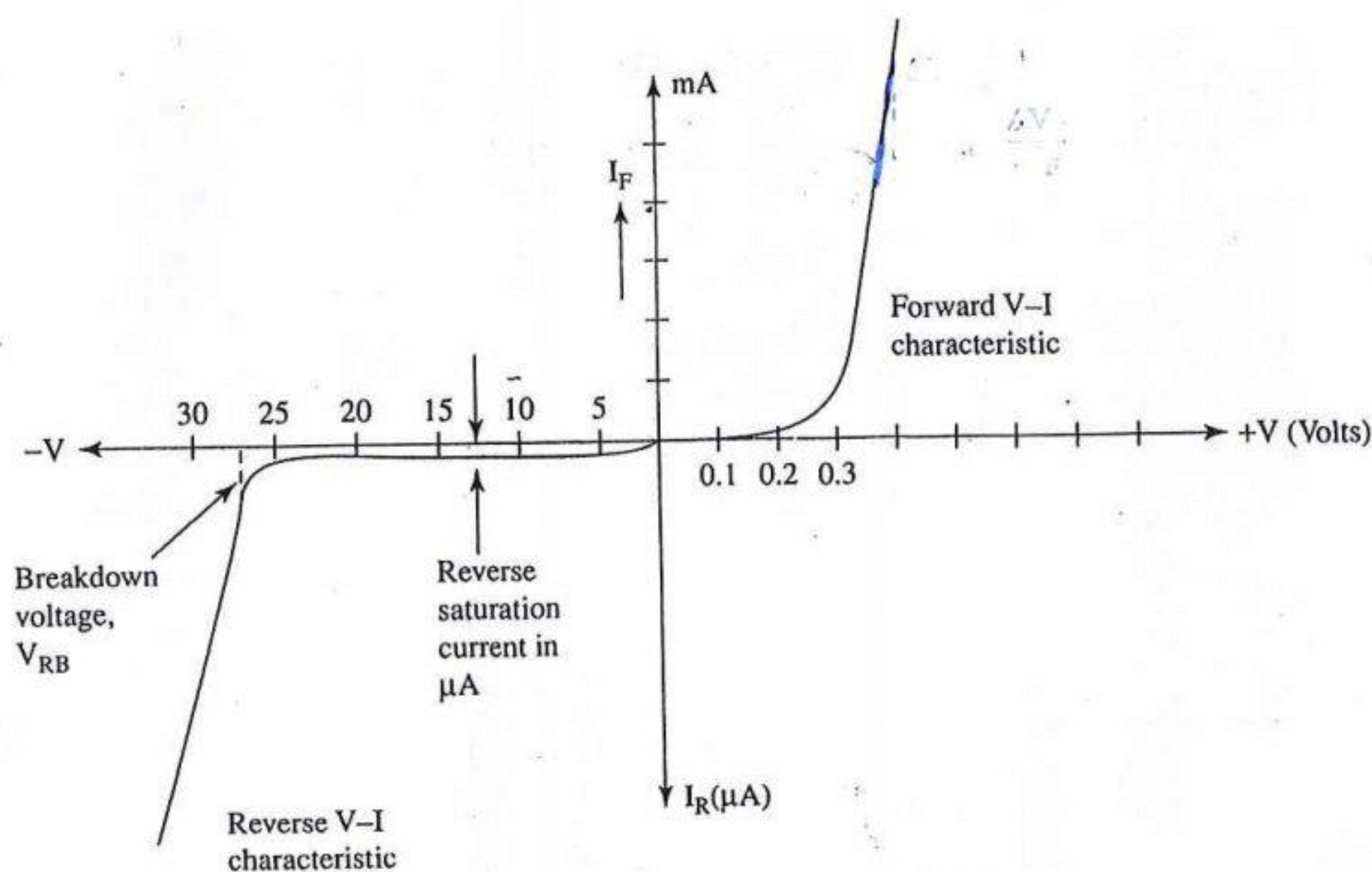


fig: V-I characteristics of reverse bias

⇒ Volt-Ampere characteristics of a Diode or V-I characteristics of a diode ⑦

- * When a p-n junction diode is connected to a source of supply such that it is forward biased, the relationship between the applied voltage V' and the current flowing I' gives us the forward V-I characteristics.
- * When the applied voltage is gradually increased, for a small value of forward voltage, the forward current is very small.
- * At a voltage near 0.3V, current suddenly increases.
- * Cut-in Voltage: The voltage at which forward current starts increasing is called the cut-in voltage of the diode.
Cut in voltage for Si is 0.7V.
Cut in voltage for Ge is 0.3V.
- * When the p-n junction is reverse biased, resistance is very high & ideally no current flows.
- * Due to minority charge carriers, a negligibly small current flows. This current is called leakage current of the diode.
- * If the reverse voltage is increased to a large value, at one stage the p-n junction breaks down & there will be a sudden increase in the reverse current.
- * Breakdown Voltage: The reverse voltage at which the diode breaks down & a large reverse current starts flowing is called the breakdown voltage.
- * At the breakdown voltage, current continues to increase.

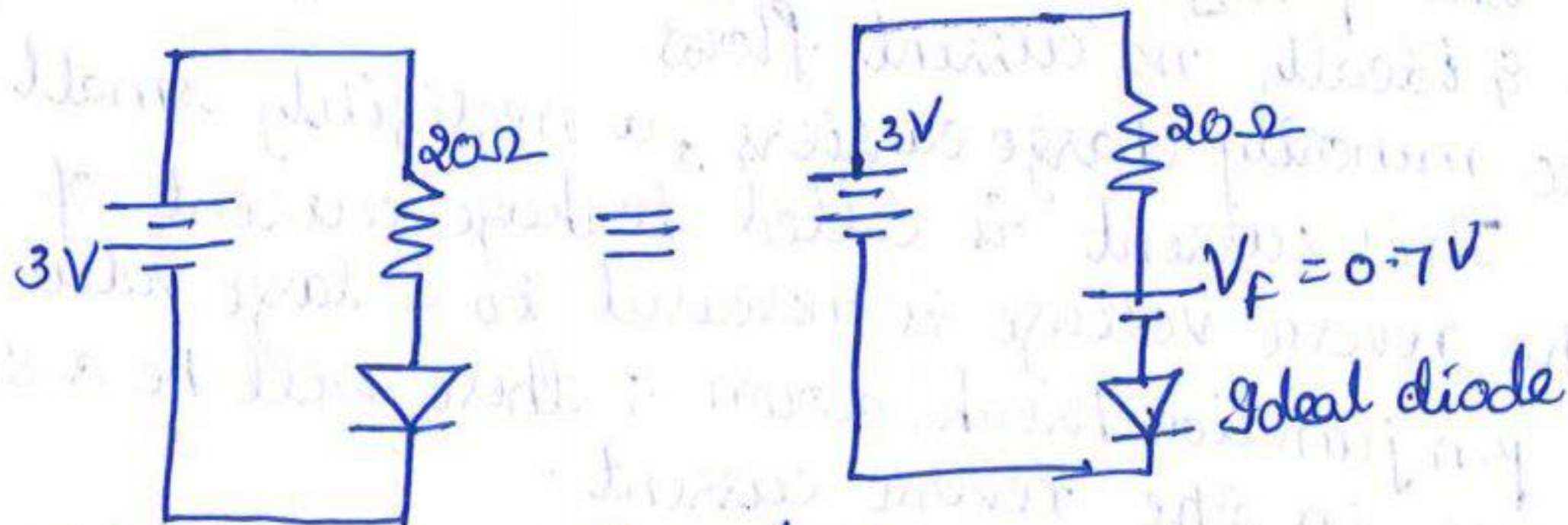


⇒ Ideal Diode:

- * An ideal diode will allow current to flow in one direction & opposes current flow in the other direction.
- * An ideal diode acts as a conductor in forward bias & acts as an insulator in reverse bias.
- * An ideal diode has zero forward resistance & infinite reverse resistance.

Problem: A Silicon diode is connected across a 3V supply with a series resistance of 20Ω . Neglecting diode resistance, calculate the diode current.

Solution: A Silicon diode has $V_F = 0.7V$



Applying Kirchhoff's voltage law,

$$3V - 20I - 0.7V = 0$$

$$20I = 3 - 0.7$$

$$20I = 2.3V$$

$$I = \frac{2.3}{20} = 0.115A$$

⇒ Diode Parameters:

A diode is specified in terms of the following parameters

1. Forward Voltage drop V_F
2. Reverse Breakdown Voltage V_{RB}
3. Reverse Saturation Current I_R
4. Dynamic Resistance r_d
5. Maximum forward current I_{FM}

⇒ Dynamic Resistance (r_d):

It is defined as the ratio of change in voltage to the change in current when AC voltage is applied.

$$r_d = \frac{\Delta V_F}{\Delta I_F}$$

⇒ Diode Ratings:

- * Diodes are available in low, medium and high current ratings.
- * Diodes of low-current ratings are used in electronic switching circuits i.e., they work as switches.
- * Medium current diodes have a maximum current rating of 400mA & reverse voltage of about 200V.
- * High current diodes are called power diodes. They are rated for high current & high reverse voltage ratings.

⇒ Zener Diode:

- * Zener diode is a p-n junction semiconductor which is operated in reverse breakdown region.
- * Zener diode is designed to be operated in the reverse biased condition.



fig: Symbol of Zener diode

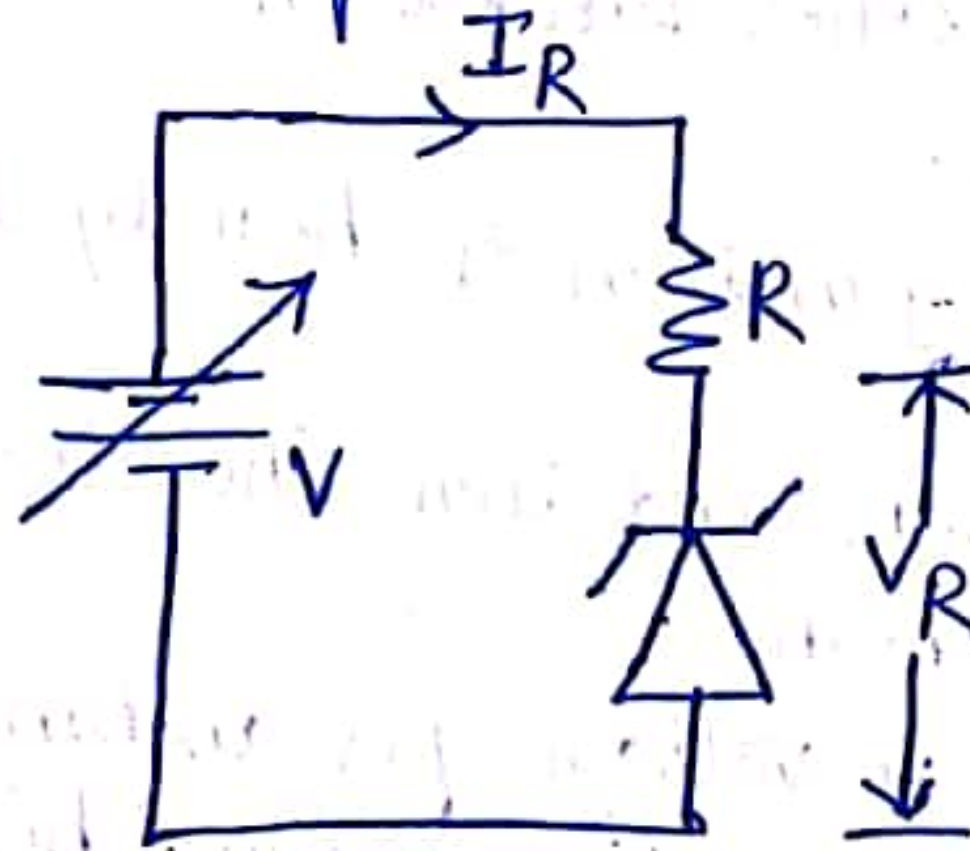


fig: Zener diode circuit

- * A diode with a very sharp breakdown voltage is called Zener diode.
- * When the diode is reverse biased, very small current, called reverse saturation current flows.
- * If the reverse voltage is increased continuously, the junction breaks down & suddenly large reverse current flows.

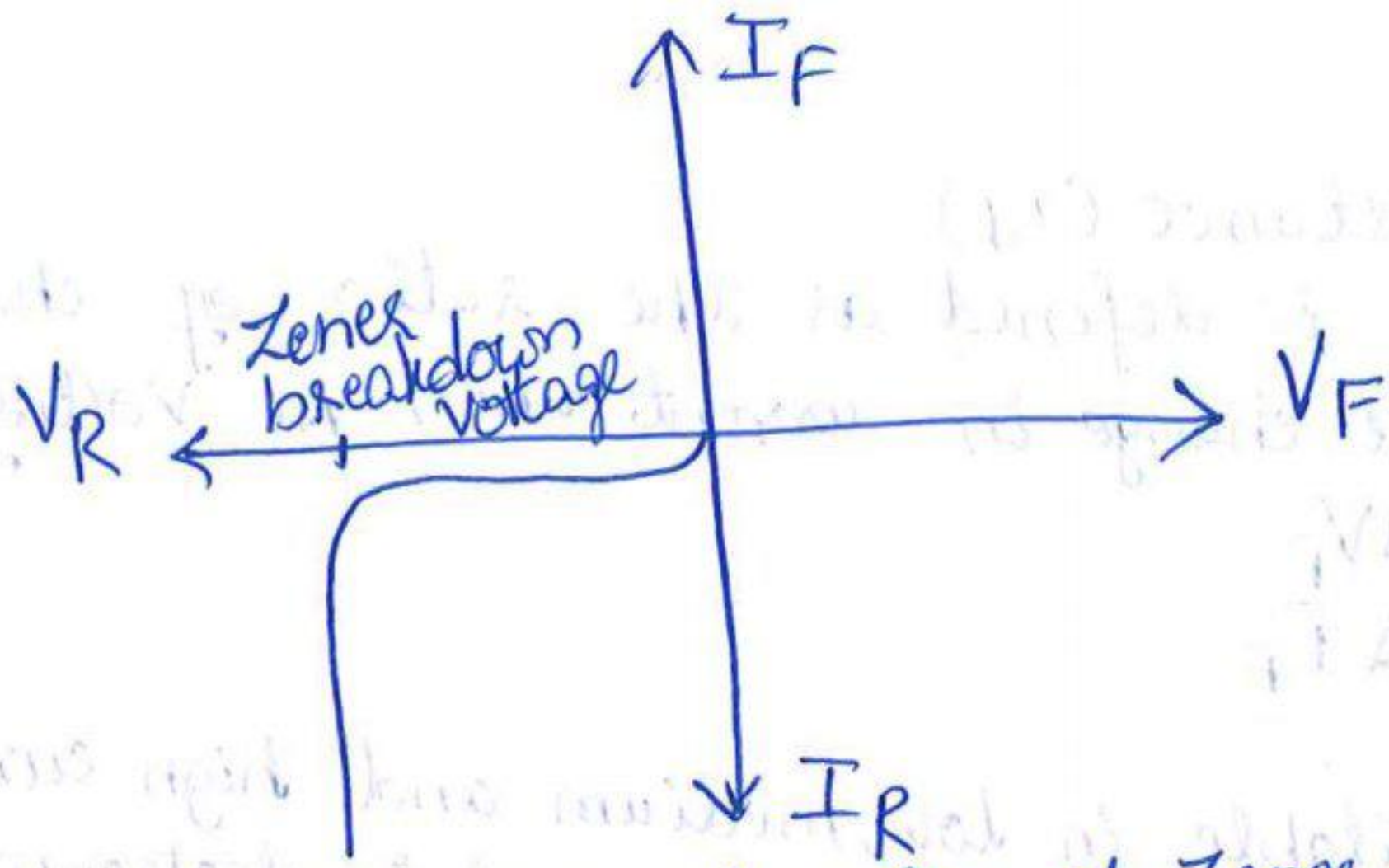


Fig: V-I characteristics of Zener diode

- * The reverse current can be controlled using a Resistor 'R' so that excess heat produced due to heavy current flow will not damage the diode.
- * Zener diode is designed such that it can be continuously operated in the breakdown region without causing any damage.

Breakdown Mechanisms:

- * There are 2 ways in which breakdown can occur in zener diode. They are:
 - ① Avalanche Breakdown
 - ② Zener Breakdown

⇒ Zener Breakdown:

- * When the p-n junction is heavily doped, the depletion region is very narrow.
- * Under reverse bias condition, the electric field across the depletion layer is very high.
- * Electric field is voltage per distance.
- * Due to narrow depletion region & high reverse voltage, the electric field is very high.
- * Strong electric field causes electrons to break away from parent atoms.
- * This kind of breakdown due to strong electric field intensity is called Zener breakdown.

* Avalanche Breakdown:

- * As the applied reverse bias voltage increases, the velocity of the minority charge carriers increases.
- * There will be collisions between these minority charge carriers & electrons involved in the covalent bonds of the crystal structure.

- * The collisions will make the electrons to break the covalent bonds.
- * Electron-hole pairs are generated very quickly and in large numbers.
- * At this stage, the junction is said to be in breakdown & current starts increasing rapidly.
- This type of breakdown is called avalanche breakdown.
- * Zener breakdown occurs at voltage less than 5V.
- Avalanche breakdown occurs at voltage higher than 5V.

⇒ Zener Resistance:

It is the dynamic resistance of the zener diode. Zener resistance is defined as the ratio of change in Zener voltage to the change in Zener current.

$$R_z = \frac{\Delta V_z}{\Delta I_z}$$

⇒ Applications of diode as Switch: A switch has 2 states ON
OFF

- * A diode is a 2 terminal device with one terminal as positive & the other as negative.
- * P-region is known as Anode & N-region is known as cathode.
- * A diode allows the flow of electricity in one direction only.
- * When the +ve terminal of the diode is connected to the +ve terminal of the battery & the -ve terminal of the diode to the -ve terminal of the battery, the diode acts as a closed switch.
- * Hence current flows through the circuit.
- * When the diode is forward biased → it is in ON state

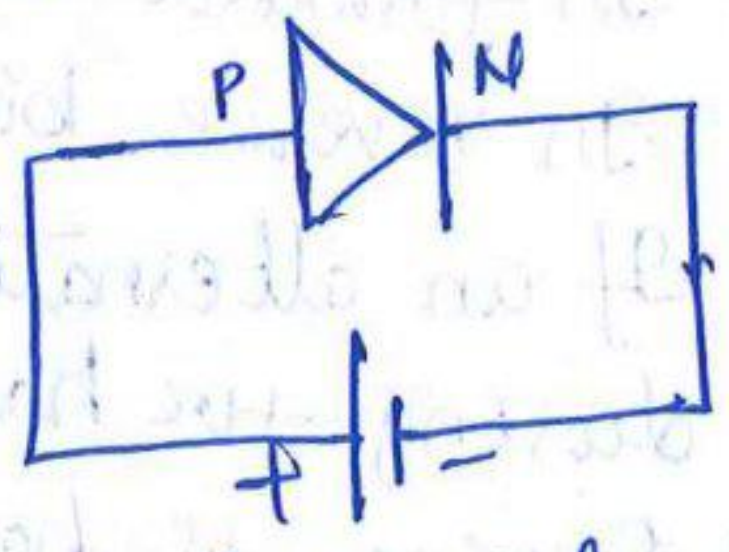
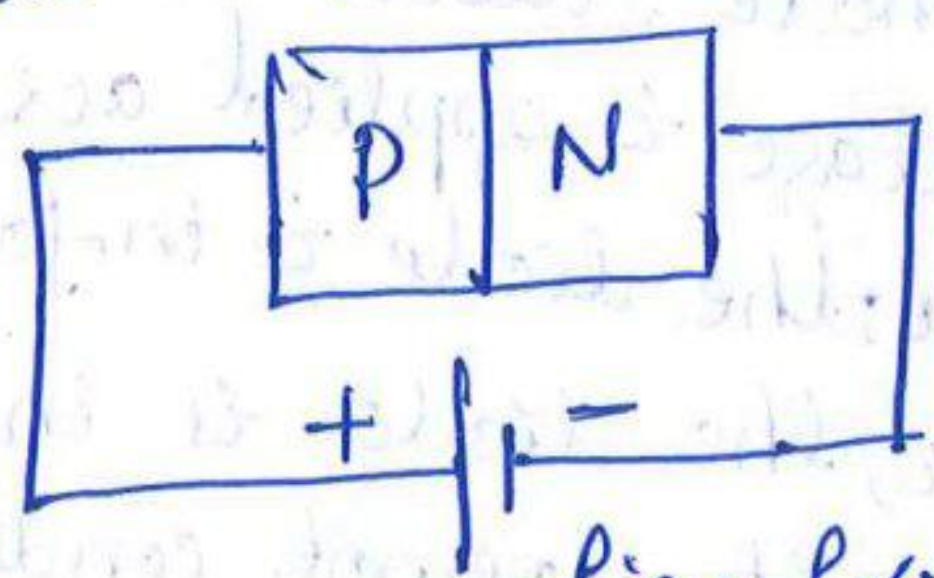


fig: forward biased condition

- * When the +ve terminal of the diode is connected to the -ve terminal of the battery and the -ve terminal of the diode is connected to the +ve terminal of the battery, the diode acts as an open switch.
- * It does not allow current to flow through it.
- * It is in reverse biased condition & the diode is in OFF state.

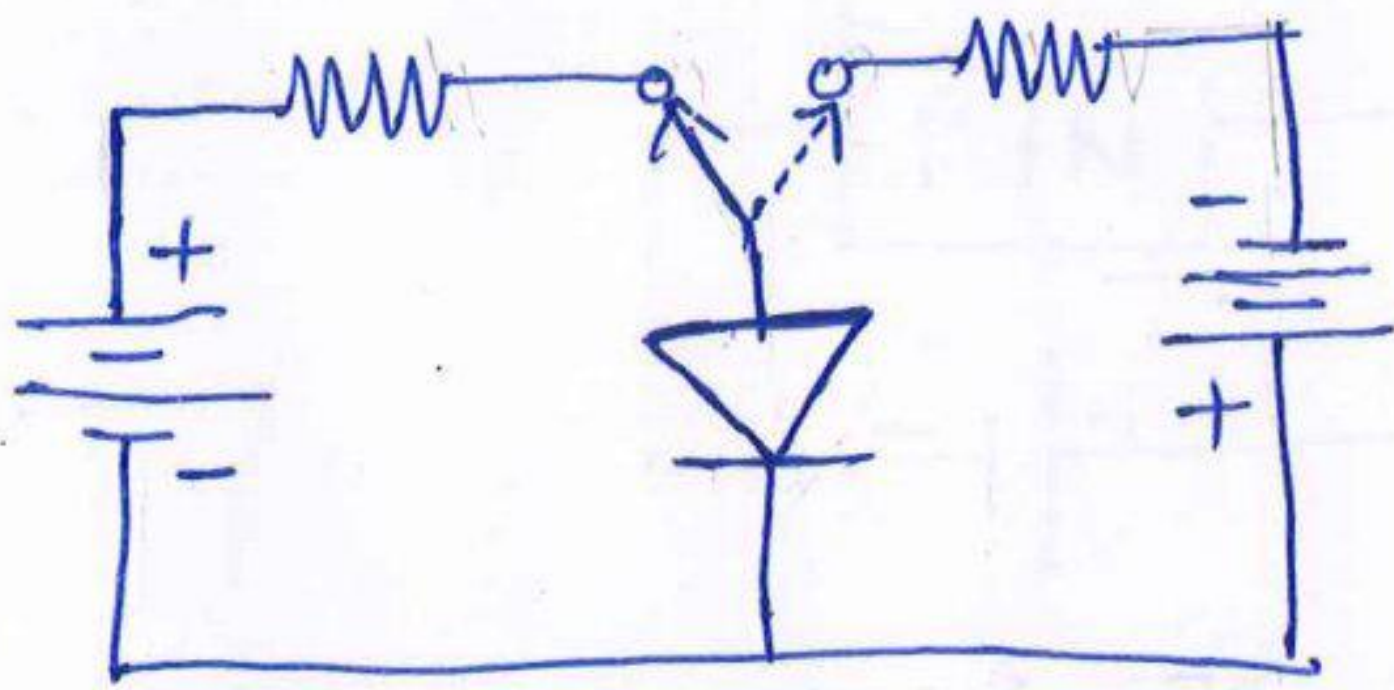


fig: Diode as a Switch

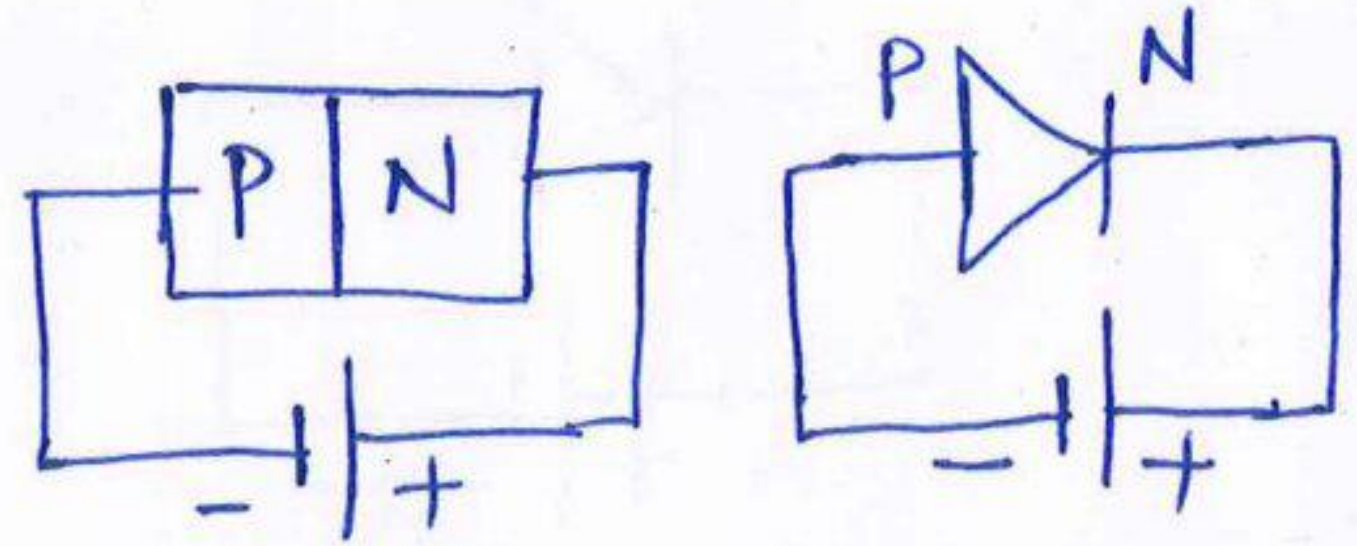


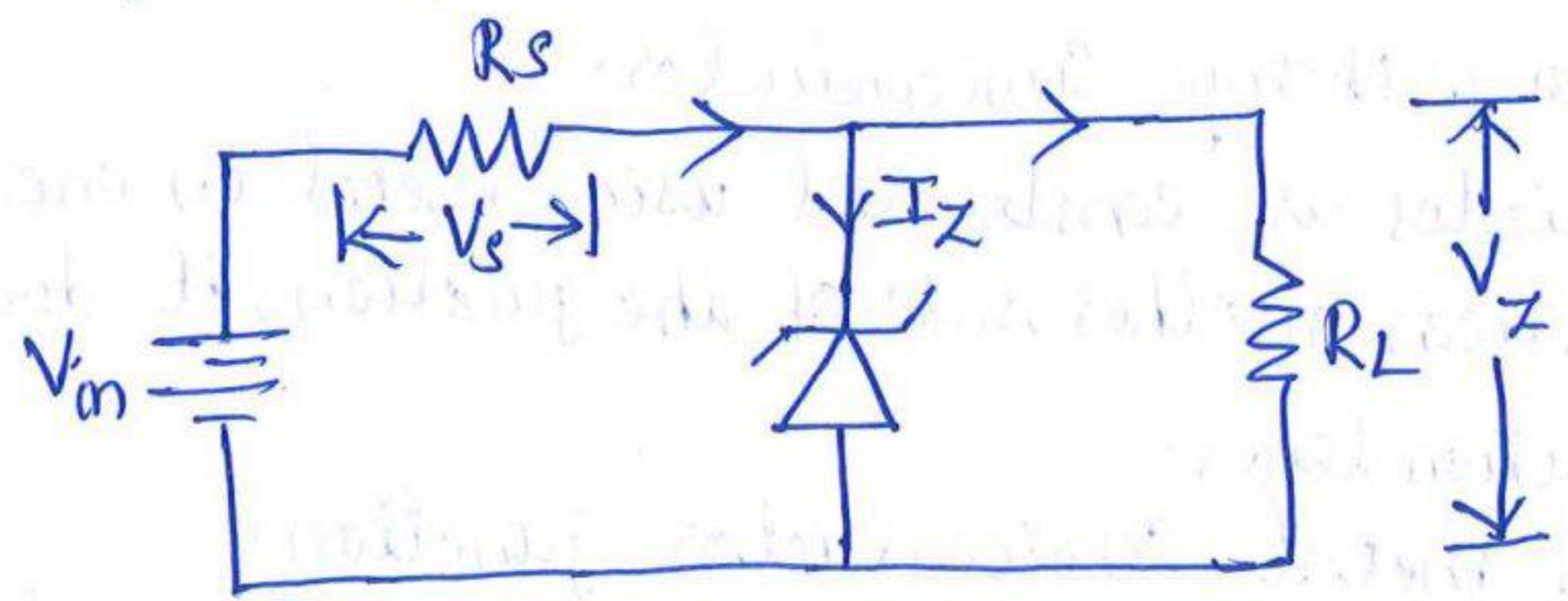
fig: Reverse biased Condition

⇒ Applications of diode as Rectifier:

- * Rectifier: It is an electronic circuit consisting of diodes which carries rectification process.
- * Rectification: It is the process of converting alternating current to direct current.
- * A rectifier is a device which converts AC signal into DC signal.
- * Diode acts as rectifying element in the rectification circuit.
- * A PN junction diode can be used as rectifying element as it offers low resistance when it is forward biased & offers high resistance when it is reverse biased.
- * In forward biased mode: diode conducts electricity.
- * In reverse biased mode: diode does not conduct.
- * If an alternating voltage is applied across the diode, during +ve half cycle, the diode is in forward bias & it conducts.
- * During -ve half cycle, the diode is in reverse bias & it does not conduct.

⇒ Zener diode as regulator :

- * Voltage Regulator is an electronic circuit which regulates the unregulated voltage.
- * The output voltage of voltage regulator is constant irrespective of the changes in the input voltage or current.



- * Zener diode is connected parallel to load R_L .
- * The voltage across the load has to be regulated and must not cross the value of V_z .
- * A suitable zener diode with a breakdown voltage near the voltage we require across the load is taken.
- * Zener diode is connected in reverse bias condition.
- * When the voltage across the diode exceeds the zener breakdown voltage, current starts flowing through the diode.
- * As the load is parallel to the diode, the voltage drop across the load is equal to the zener breakdown voltage.
- * The zener diode provides a path for the current to flow & hence the load gets protected from excessive currents.
- * Hence constant output is maintained.



⇒ Special Purpose Diodes:

⇒ Schottky Diode:

⇒ Schottky diode is a metal semiconductor diode with a low forward voltage drop & fast switching speed.

⇒ Schottky diodes are constructed using a metal electrode bonded to a N-type Semiconductor.

* As these diodes are constructed using metal on one side & doped silicon on other side of the junction, it does not have depletion layer.

* It has a metal-semiconductor junction.

* The width of the junction depends upon the type of metal & semiconductor used.

* When it is forward biased, electrons move from the N-type material to the metal electrode allowing current to flow.

⇒ Current in the schottky diode is due to the majority carriers.

⇒ When it is reverse biased, current conduction stops very quickly.

⇒ For a schottky diode there is a very rapid response to changes in bias.

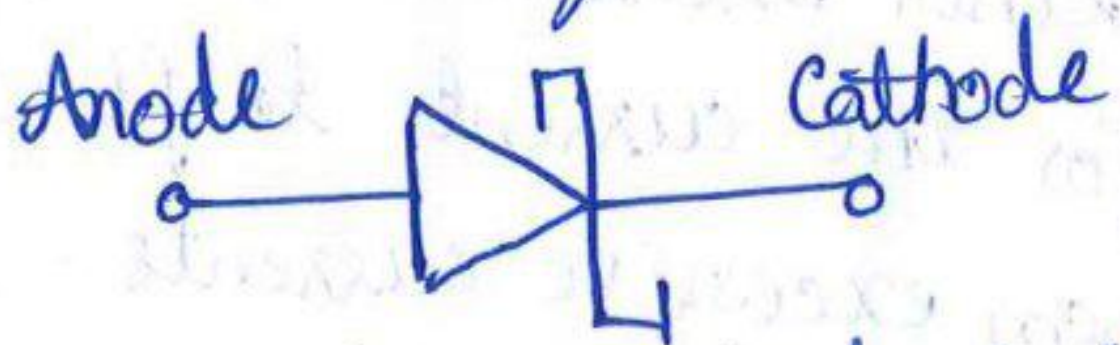


fig: Symbol of Schottky diode Junction

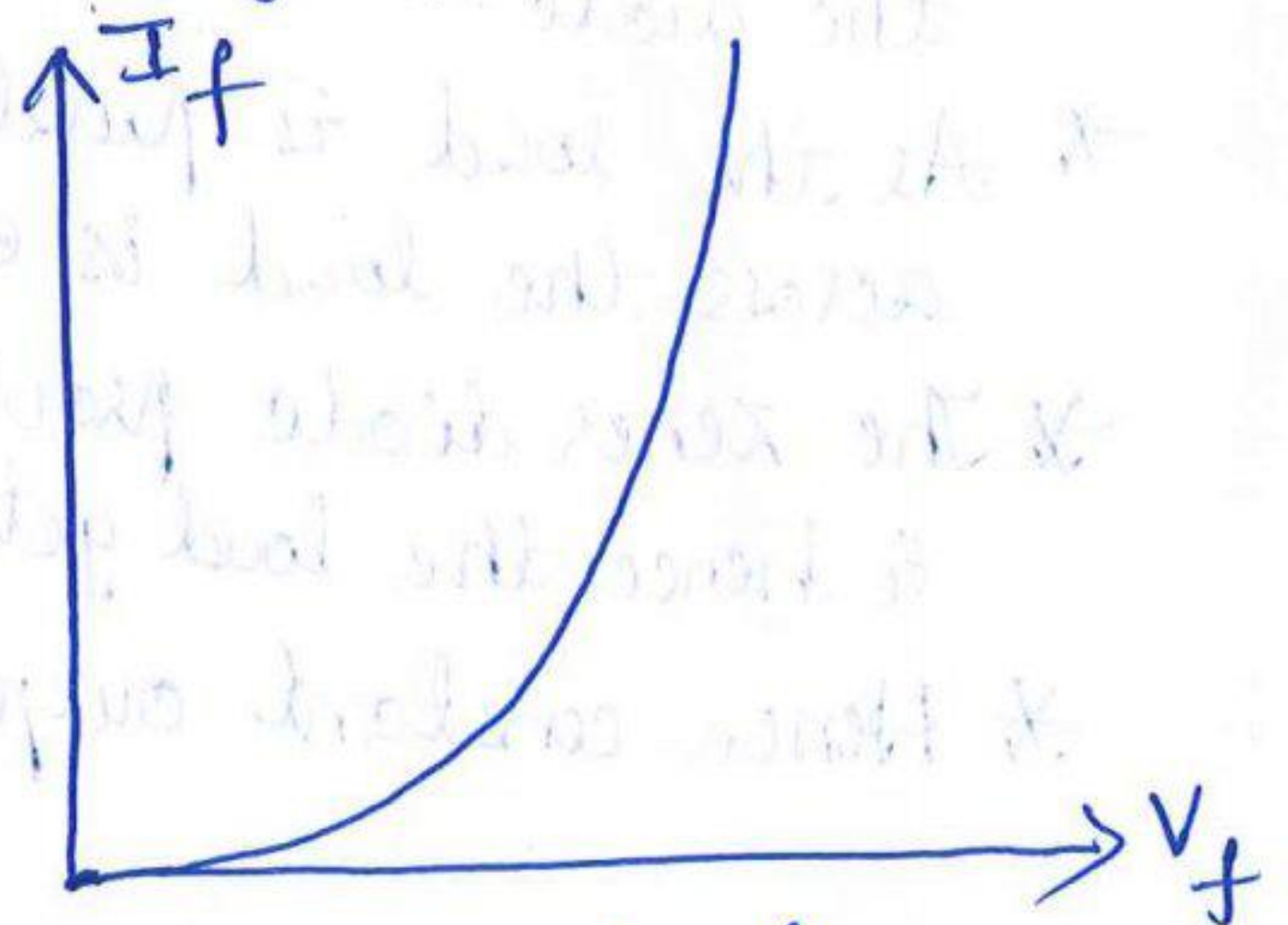
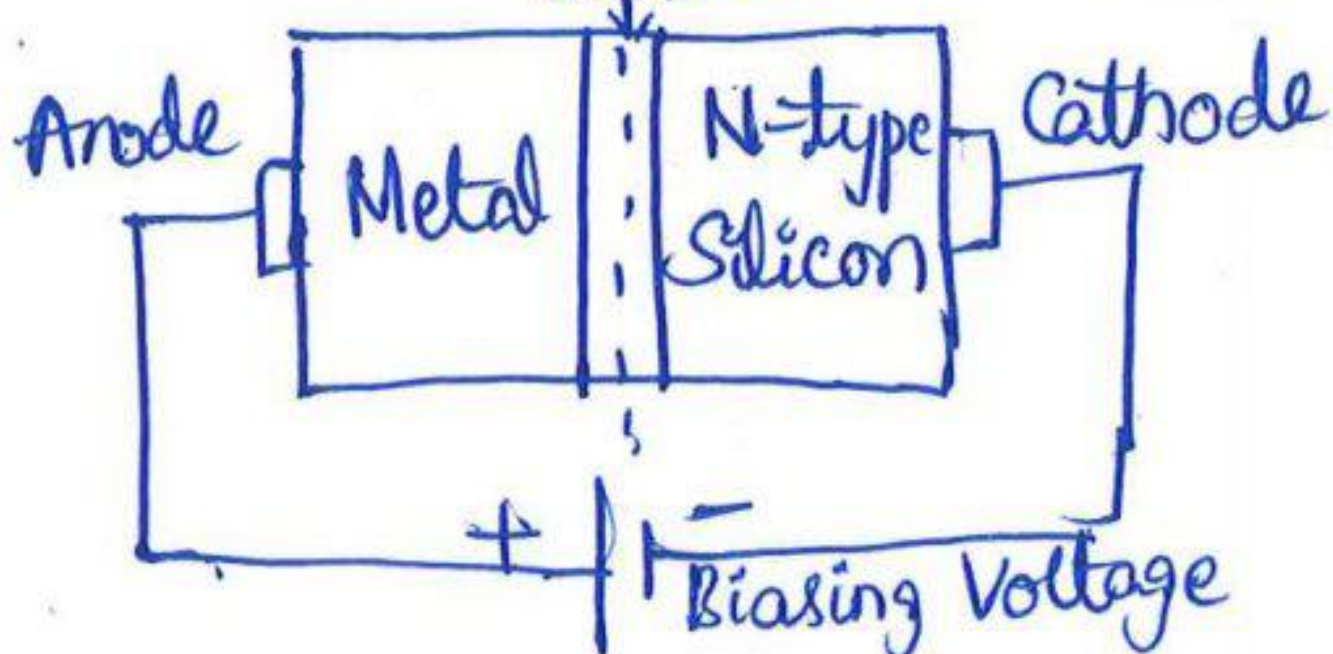


fig: V-I characteristics

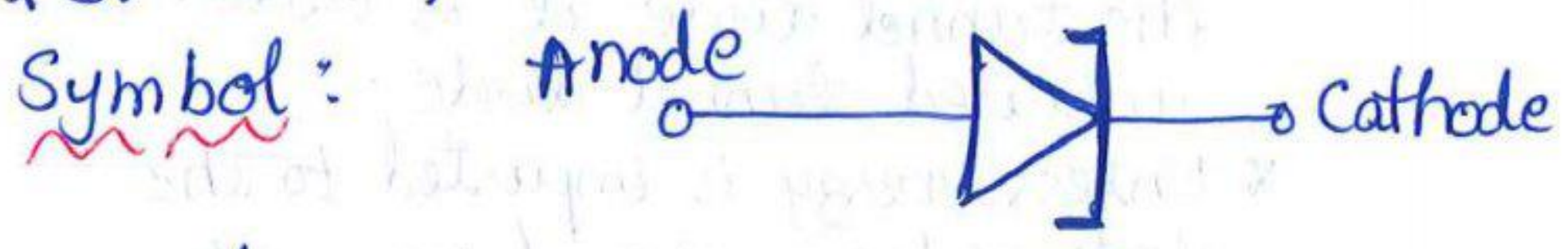
⇒ Tunnel Diode :

Definition : A thin junction diode which exhibits negative resistance under low forward bias condition is known as tunnel diode.

- * Tunnel diode is highly conductive.
- * It is a heavily doped pn junction diode.
- * It produces current because of tunnelling.

Tunnelling : It is the phenomenon of conduction in semiconductor material in which charge carrier punches the barrier instead of climbing through it.

- * It is a fast switching device.



Operation :

- * When the diode is forward biased, due to heavy doping, immediate conduction takes place.
- * Current in the diode reaches maximum value I_p (peak current), when voltage V_p (peak voltage) is applied across it.
- * When the voltage further increases, current across the terminals drops & continues to drop till point B is reached.
- * The region between A to B is called the -ve resistance of tunnel diode.
- * At point B, current increases as applied voltage increases.
- * Negative resistance region is the region in which increase in the voltage results in decrease in current.

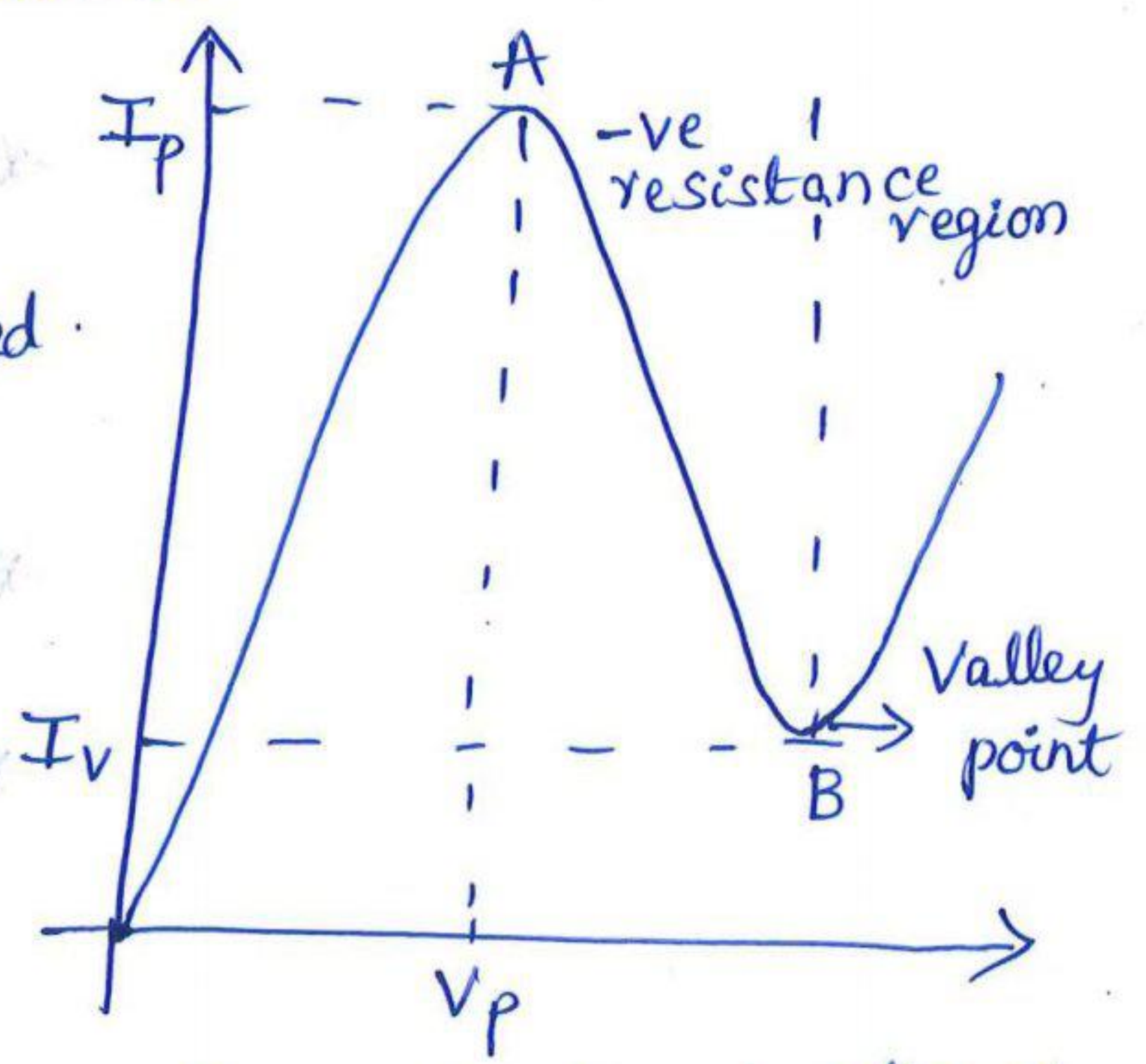
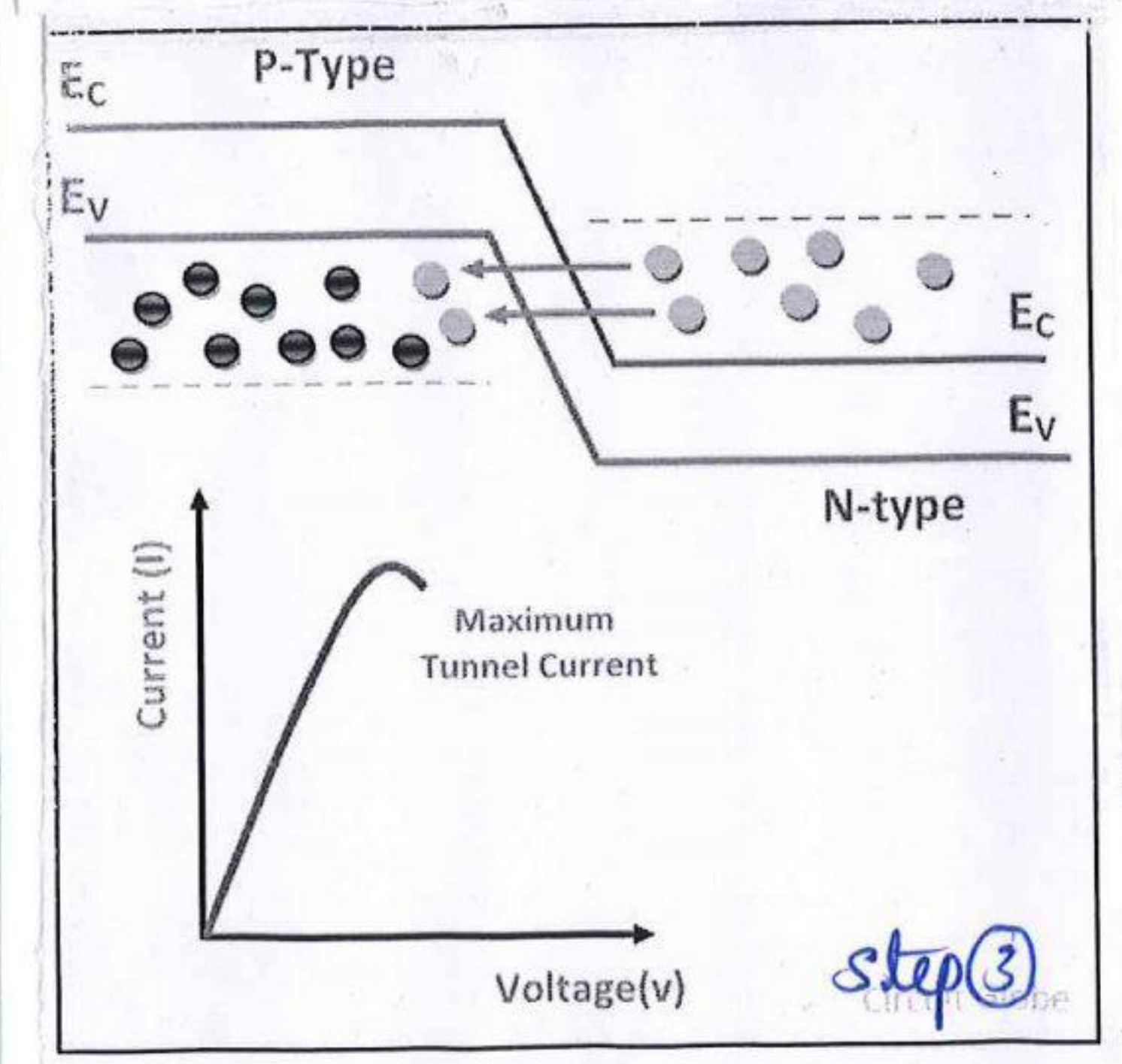
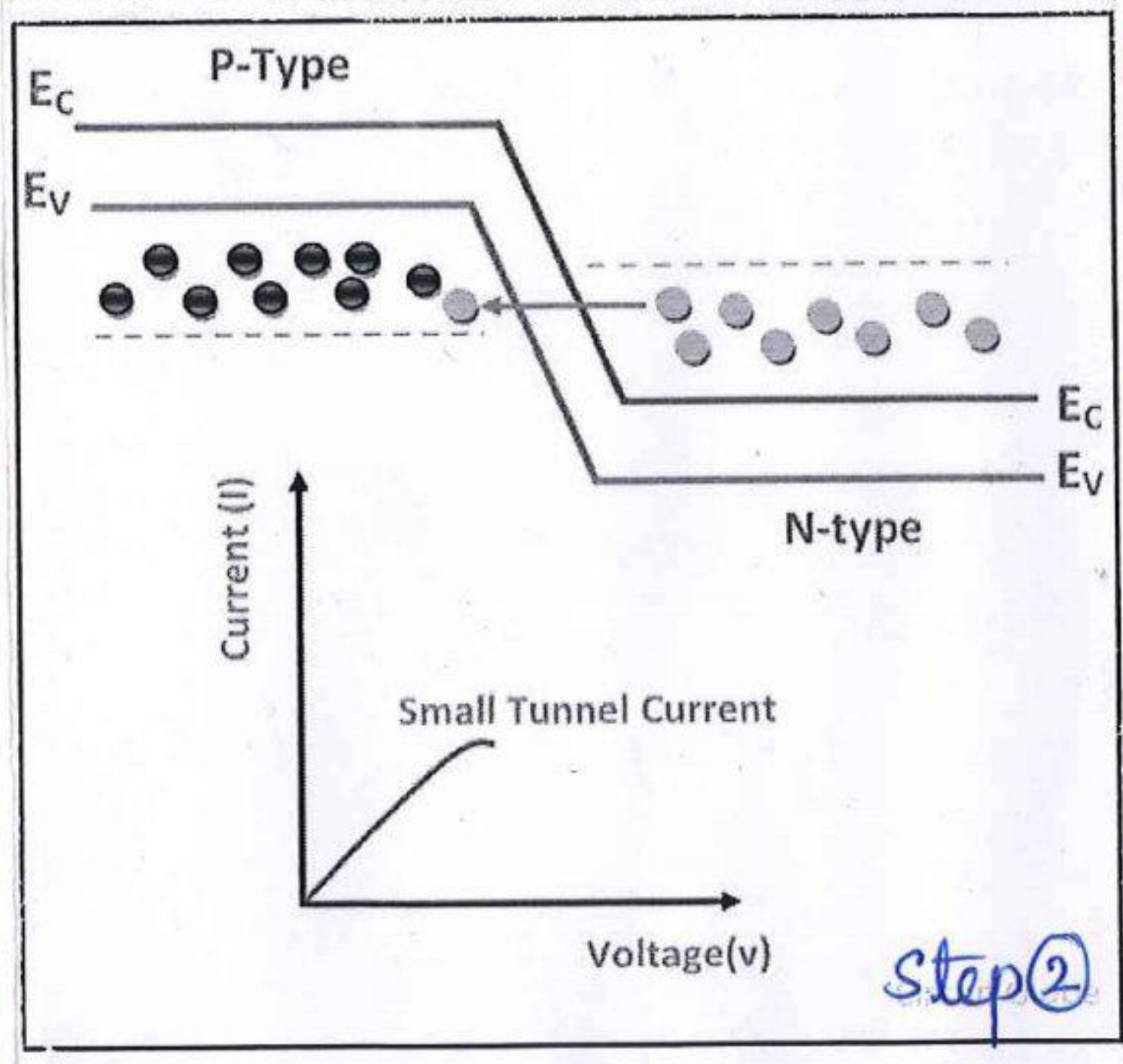
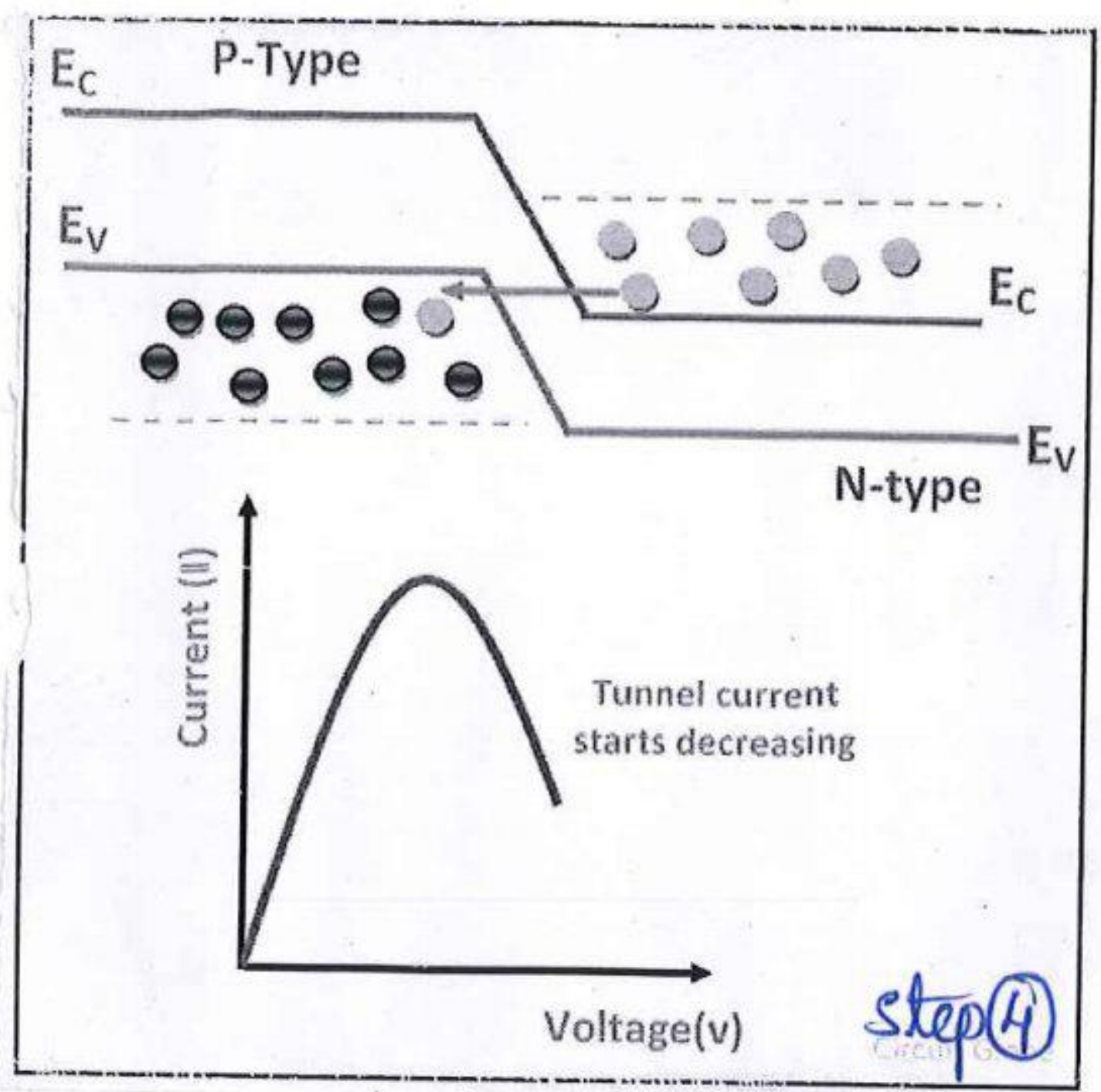
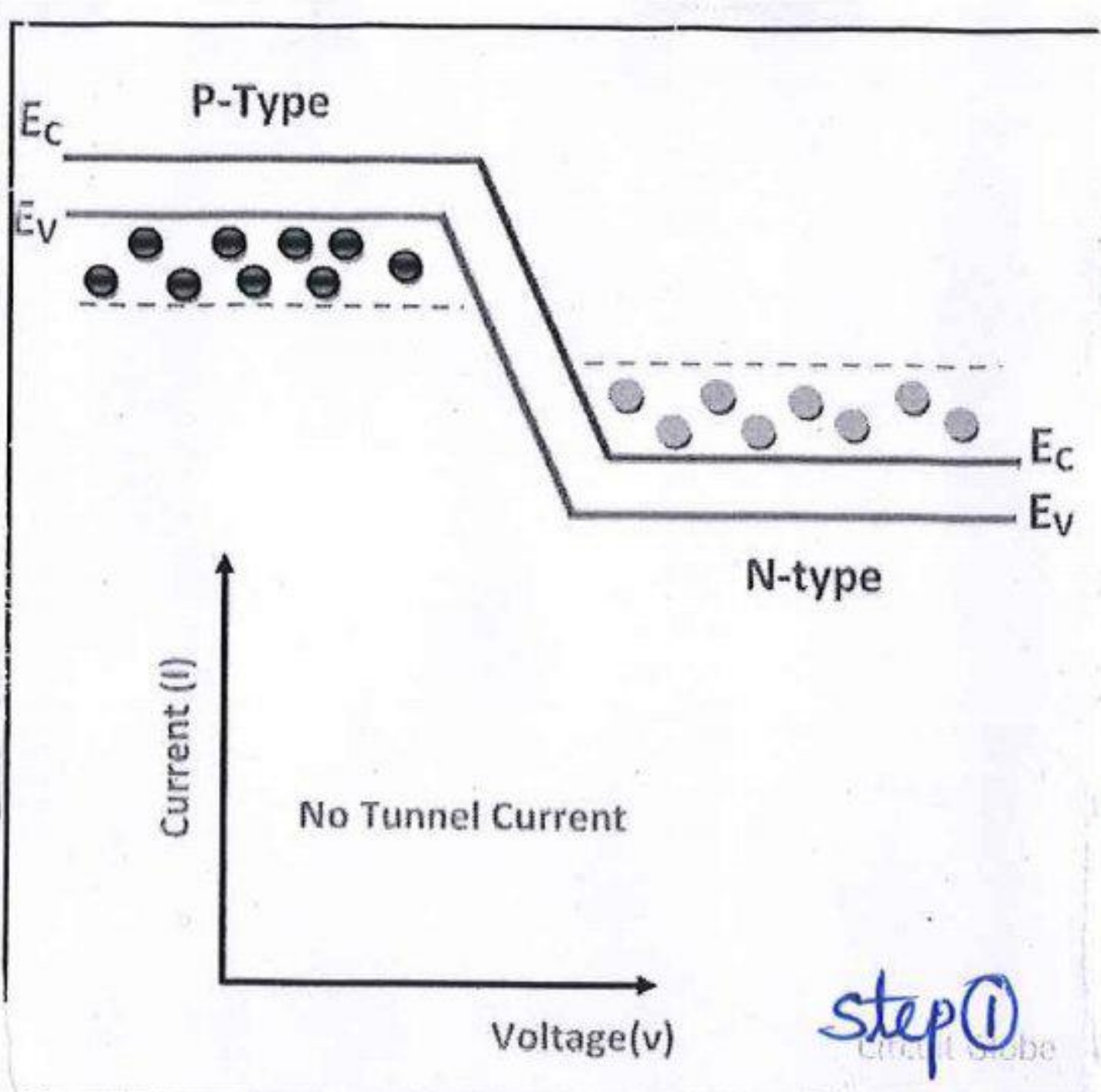


fig: V-I characteristics of tunnel diode



Step 1: Unbiased tunnel diode

- * When no voltage is applied to the tunnel diode it is called unbiased tunnel diode.
- * Unless energy is imparted to the electrons from external source, they cannot cross the barrier.
- * Forward current is zero.

Step 2: When small voltage is applied to the diode.

- * In this case, small number of electrons in the conduction band of N-region will tunnel to the valence band of P-region.
- * This creates small forward current.
- * This tunnel current flows with small application of voltage.

Step 3 : Applied voltage is slightly increased

- * When the voltage applied to the tunnel diode is slightly increased, a large number of free electrons move from n-side to p-side.
- * Overlapping of conduction & valence band increases.
- * Energy level of n-side conduction band is equal to the energy level of p-side valence band, hence maximum current flows.

Step 4 : Applied voltage is further increased

- * A slight misalign of conduction & valence band takes place - current starts decreasing.

Step 5 : When the applied voltage is largely increased, conduction & valence band do not overlap & the tunnel diode operates as normal pn junction diode.

Advantages :

- * Long life
- * High Speed operation
- * Low Noise
- * Low power consumption

Applications :

- * Used in FM receivers.
- * Used in amplifiers.
- * Used in High frequency oscillators.
- * Used as a high speed switch.
- * Used in computers.

⇒ Varactor Diode:

Definition: It is defined as the diode whose internal capacitance varies with the variation of reverse voltage.

Symbol:

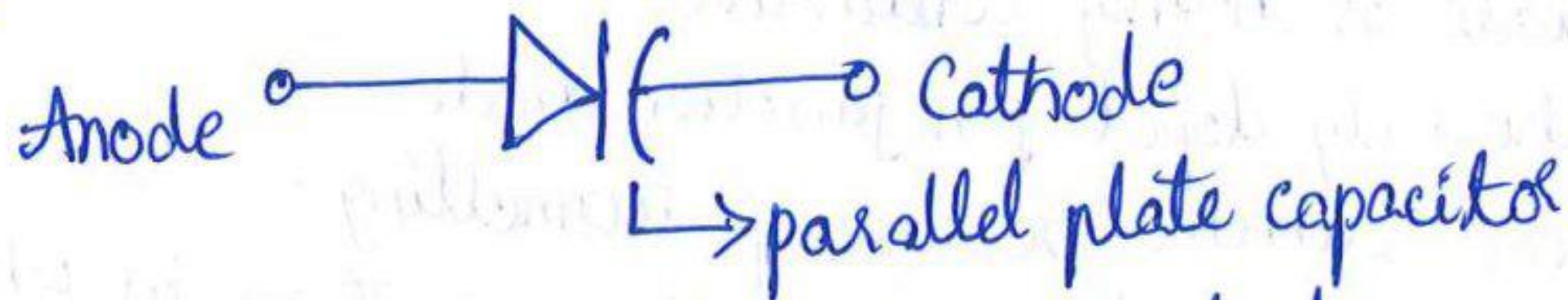


fig: Symbol of Varactor diode

* It is used for storing the charge.

* It always works in reverse bias.

* Variable diode is called voltage dependent semiconductor device as the o/p of the diode depends on i/p voltage.

* The symbol of the varactor diode has a diode & two parallel lines which represent 2 conductive plates of capacitor.

* Varactor diode is also called Varicap, voltage variable capacitance or tuning diode.

Characteristics of Varactor diode:

* As the reverse voltage V_R that is applied to the diode increases, the width of the depletion region increases & transition capacitance C_T decreases.

* Capacitance can be controlled by the applied voltage.

Applications of Varactor Diode:

1. Tuned Circuits

2. FM Modulators

3. Television Receivers

4. Automatic Frequency Control Circuits

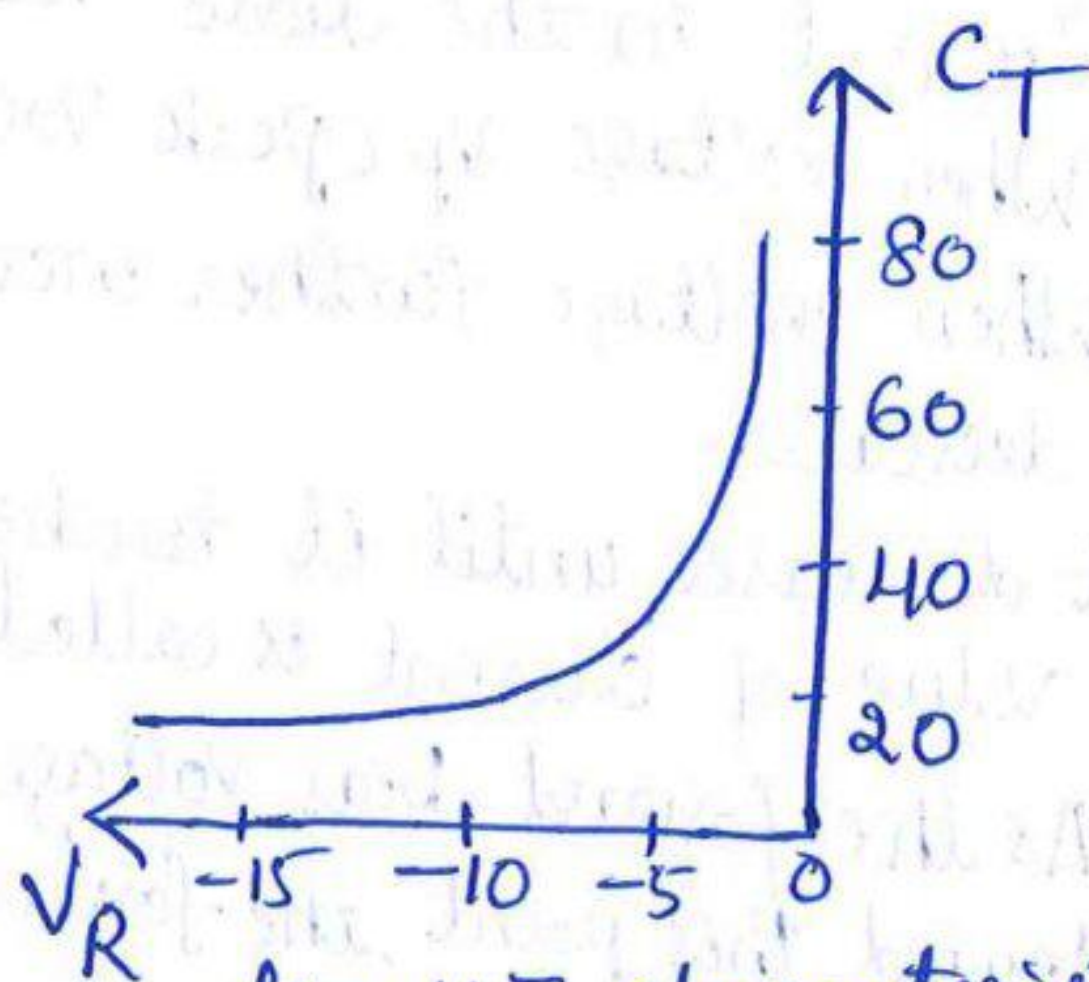


fig: V-I characteristics

⇒ Photo Diode :

Definition: It is defined as the PN junction diode that generates current when it is exposed to light.

- A photo diode converts light energy into electrical energy.
- Photo diode operates in reverse biased mode.
- Current is generated when photons are absorbed in the photo diode.

Symbol:

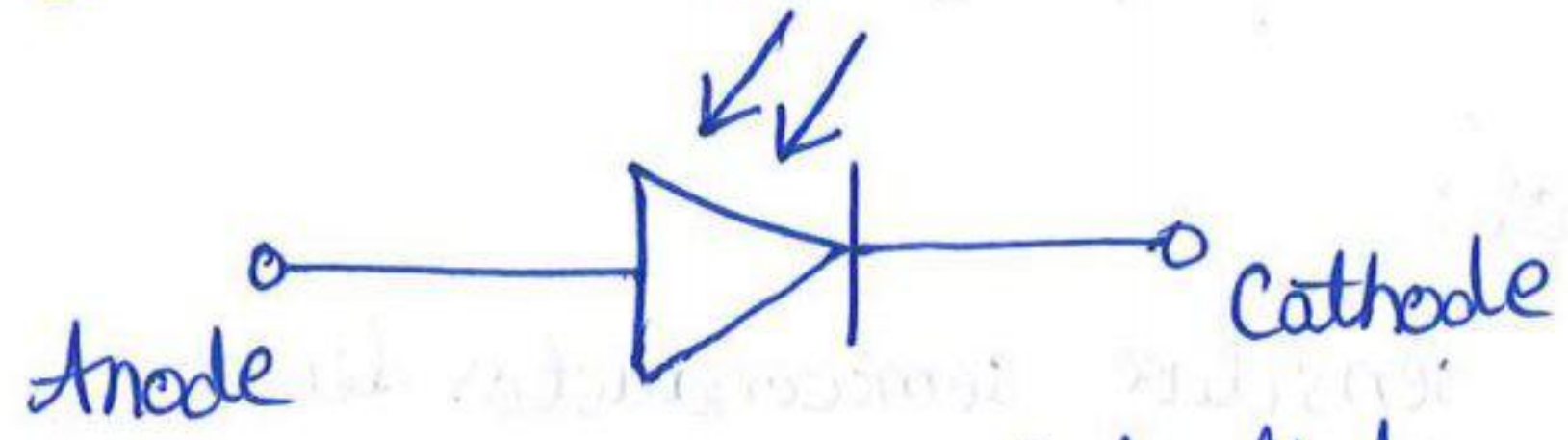
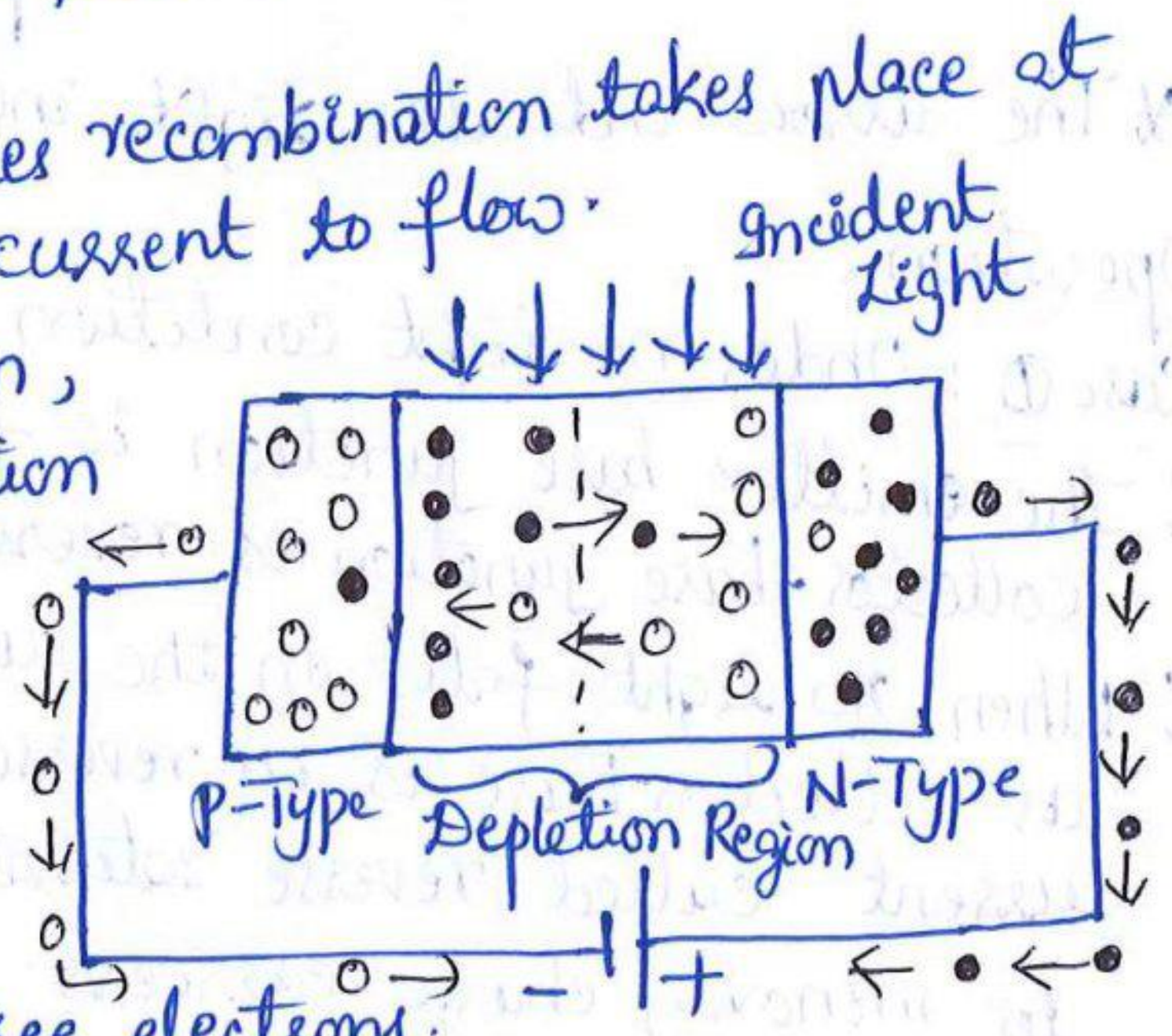


fig: Symbol of photo diode.

The arrow striking the diode represents light

Working Principle:

- The pn junction of the device is placed inside the glass material to allow light to pass through it.
- When there is no light, the reverse biased photo diode carries a very small current called dark current due to reverse voltage.
- When reverse voltage is applied to the diode, minority charge carriers experience repulsive force from the +ve and -ve terminals of the battery.
- Due to this electrons & holes recombination takes place at the junction & causes dark current to flow.
- As light falls on the junction, the temperature of the junction increases.
- Then the valence electrons gain enough energy to break the bonding with parent atoms & they will become free electrons.
- Free electrons move freely & carry electric current.



* As the intensity of light increases, more carriers are generated which causes flow of large current through the device.

Advantages :

1. Can be used as variable resistance device
2. Highly sensitive to light
3. Speed of operation is very high.

⇒ Phototransistor :

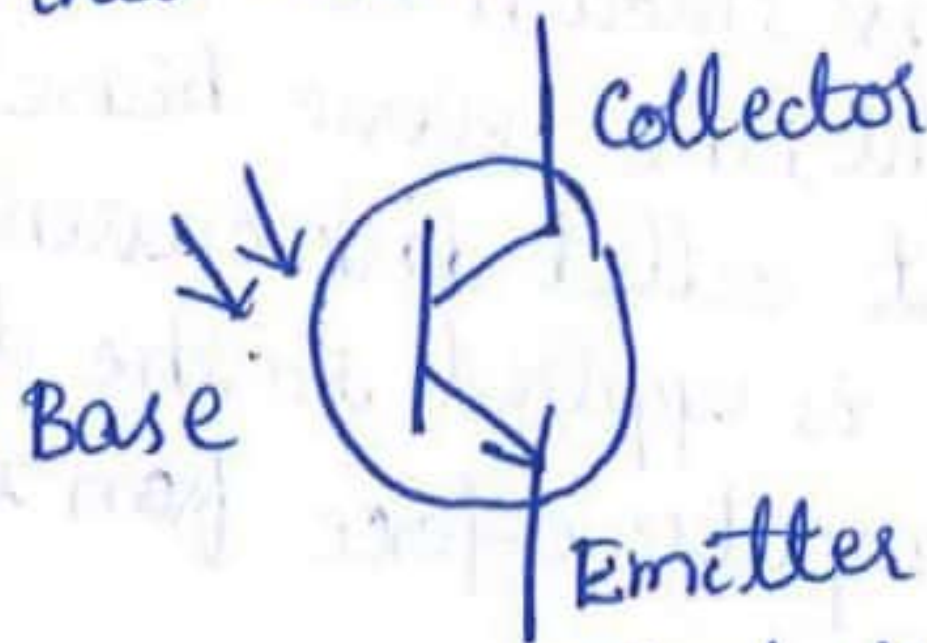
* It is a light sensitive semiconductor device.

* It converts light energy into electrical energy.

* Photons (light particles) activate phototransistor.

Definition : Phototransistor is a three layer semiconductor device which has light sensitive base region. The base senses the light & converts it into current.

Symbol :



* The arrow indicates light incident on the base.

Operation :

Case ① : Under no light condition

* The emitter-base junction is forward biased & the collector-base junction is reverse biased.

* When no light falls on the surface of the transistor, as collector-base is in reverse bias, a small amount of current called reverse saturation current flows due to minority charge carriers.

Case 2: Under light condition

- * When light enters into the base region of photo transistor, it generates electron-hole pairs.
- * The movement of electrons under the influence of electric field causes the current in the base region.
- * The brightness of light shining on photo transistor's base terminal determines the amount of current that flows.
- * Brighter light results in more current.
- * Less bright light results in less current.

Applications:

1. Used in smoke detectors.
2. Used in infrared receivers.
3. Used in CD players.
4. Used in astronomy, night vision etc.

Advantages:

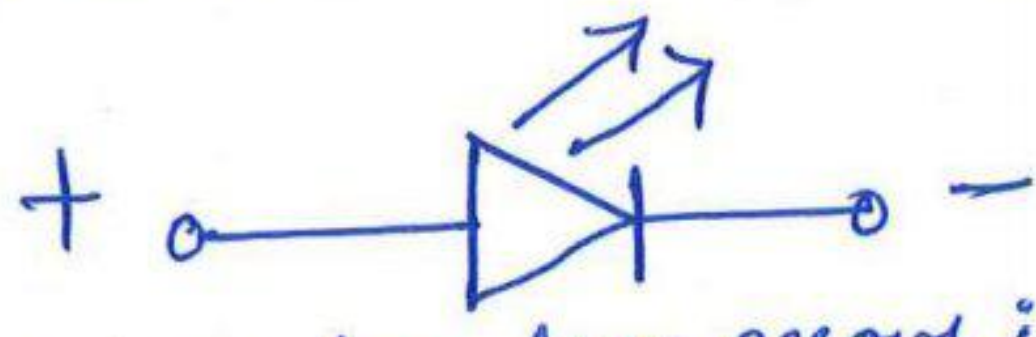
- * They produce higher current.
- * Very fast
- * Output is instantaneous.
- * Inexpensive, simple & small in size.

⇒ Light Emitting Diode (LED):

Definition:

LED is a p-n junction device, which emits light when in forward biased condition.

Symbol:



- * The symbol of LED has two arrows indicating that the device emits light energy.
- * LED converts electrical energy into light energy.
- * LED emits visible or invisible light when it is forward biased.

Basic Operation:

- * In forward bias mode, the free electrons in the conduction band recombine with the holes in the valence band.
- * During this jump from conduction band to valence band, electron emits energy in the form of photon.
- * The difference of energy between the two bands is radiated in the form of light.

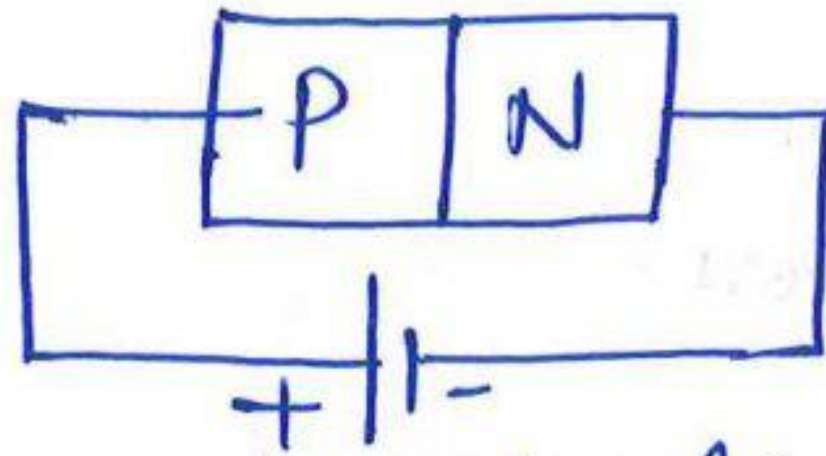
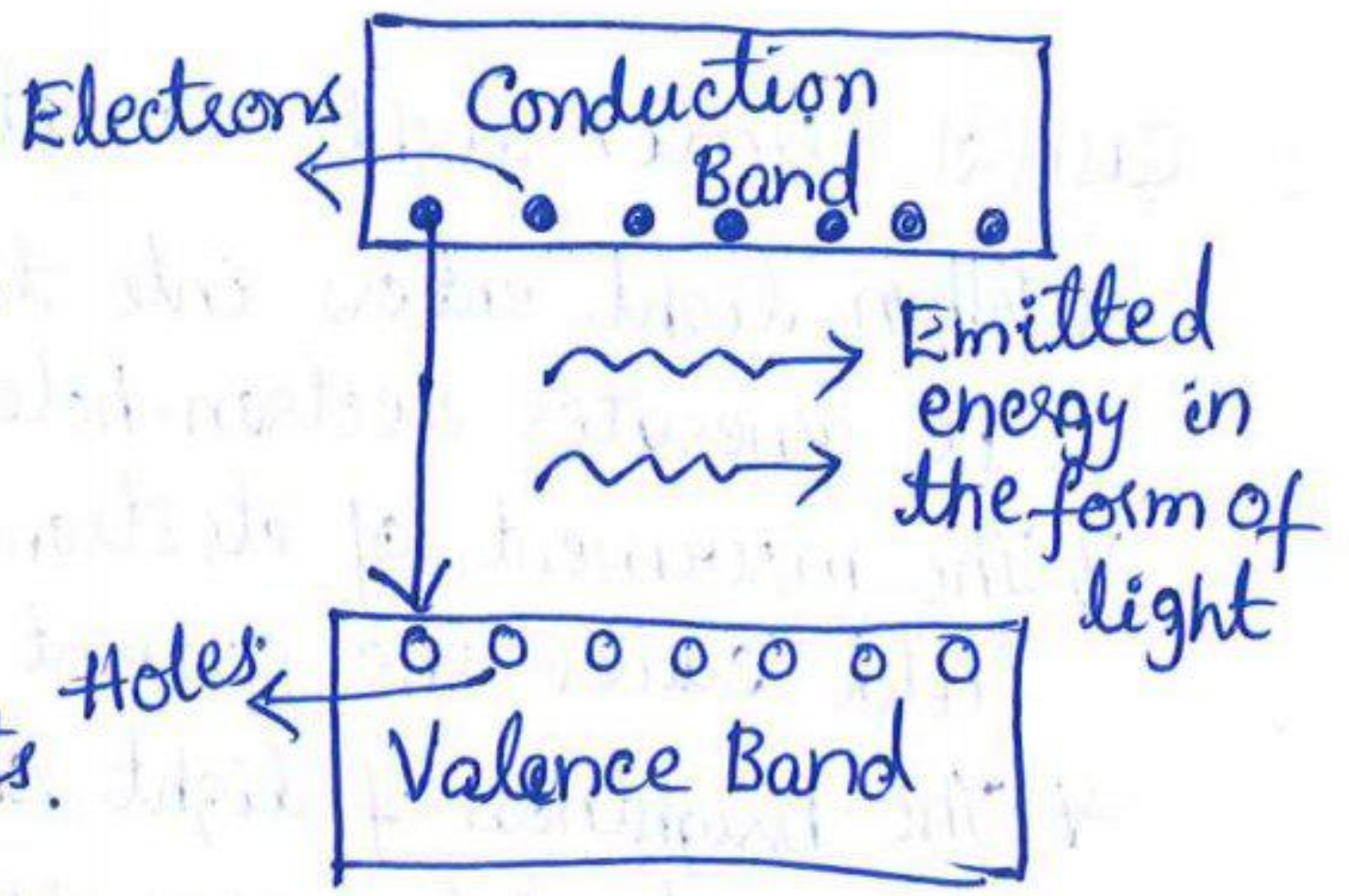
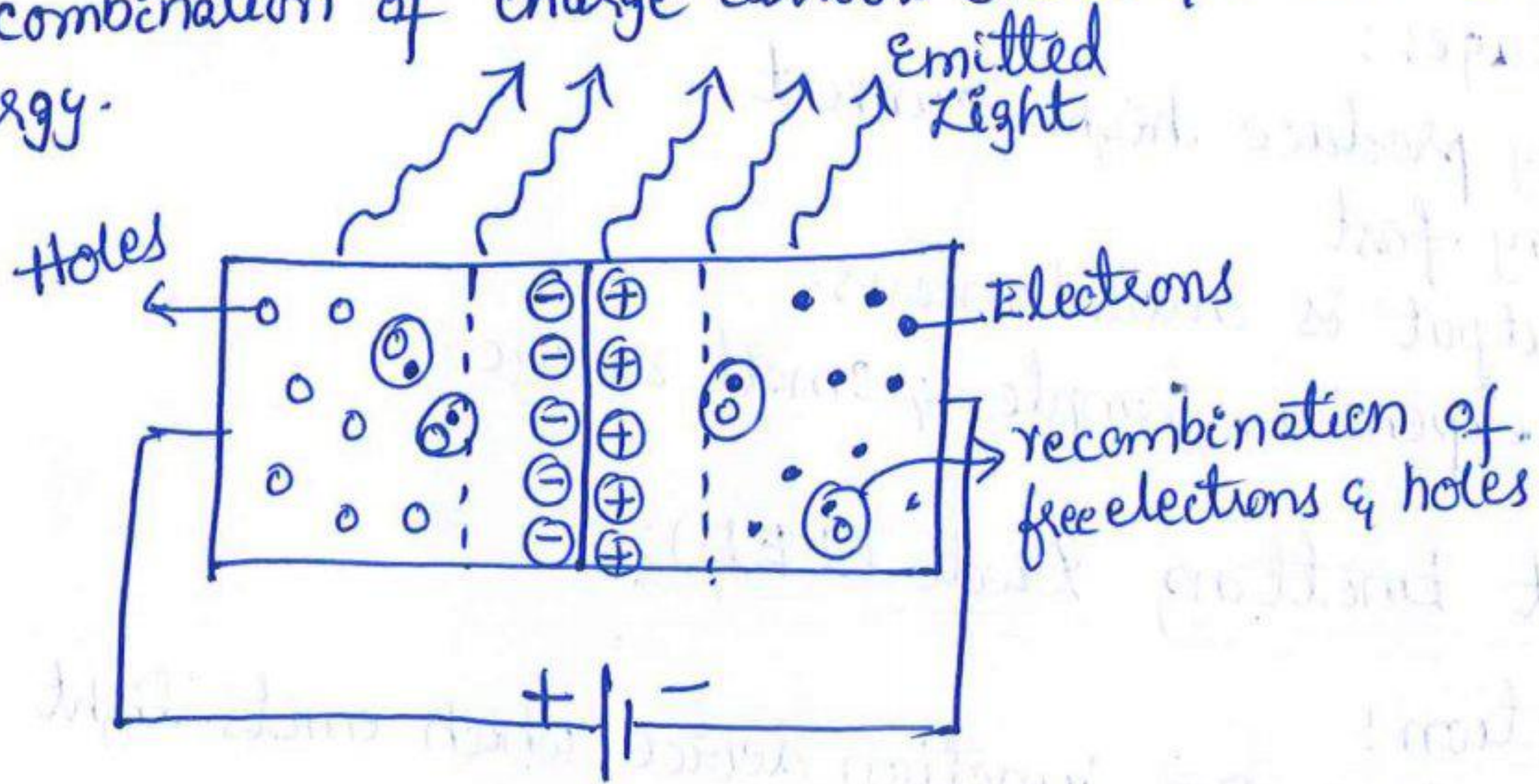


Fig: LED under forward bias

- * Each recombination of charge carriers (holes & electrons) emits some energy.



Applications:

- * Used in display devices like digital watches
- * 7-segment display
- * Burglar Alarms
- * Calculators
- * Microprocessors
- * Digital Computers
- * Optical Communications

⇒ BJT (Bipolar Junction Transistor) :

Definition of Transistor:

* Transistor = Transfer Resistor

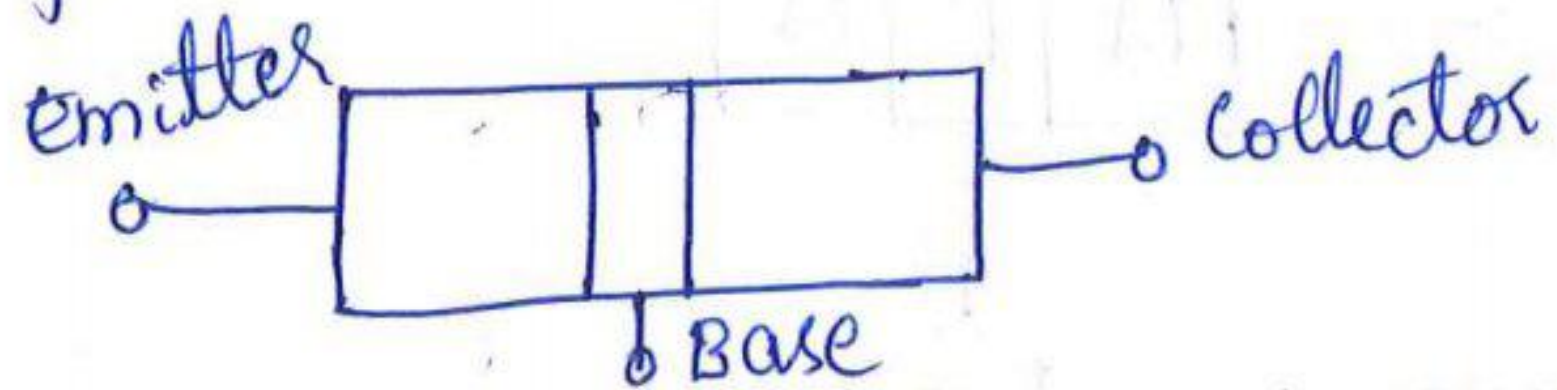
* Signal is transferred from a region of low resistance to a region of high resistance.

* A transistor is a 3 terminal semiconductor device used for amplification & switching of electrical signals.

Terminals of a Transistor:

* The three terminals of a transistor are:

1. Emitter
2. Base
3. Collector



* Emitter is heavily doped. Its main function is to supply charge carriers to base.

* Base is lightly doped. It is very thin. It passes most of the charge carriers from Emitter to Collector.

* Collector is moderately doped. Its main function is to collect majority charge carriers coming from Emitter.

Types of Transistors: Based on the type of conduction, transistors are classified as

1. Unipolar Transistor: Current conduction is due to majority charge carriers. Ex: UJT, FET

2. Bipolar Transistor: Current conduction is due to both majority & minority charge carriers.

Ex: BJT

Types of Bipolar Junction Transistor:

They are classified as

1. NPN Transistor &
2. PNP Transistor

⇒ Construction of NPN Transistor:

* It is formed by sandwiching a single P-region between two N-regions.

* Two PN junctions are connected back to back.

Emitter is of N-Type

Base is of P-Type

Collector is of N-Type

⇒ Symbol of NPN Transistor:

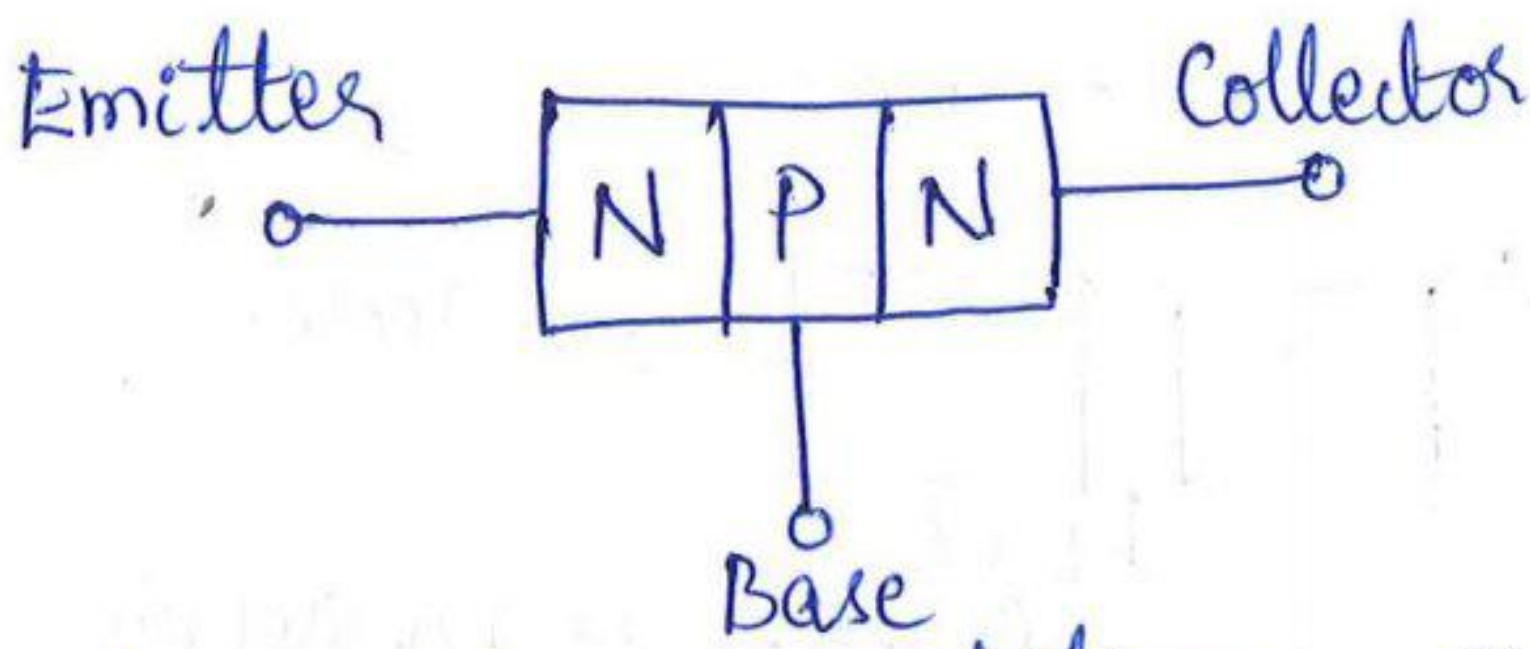


fig: Block representation

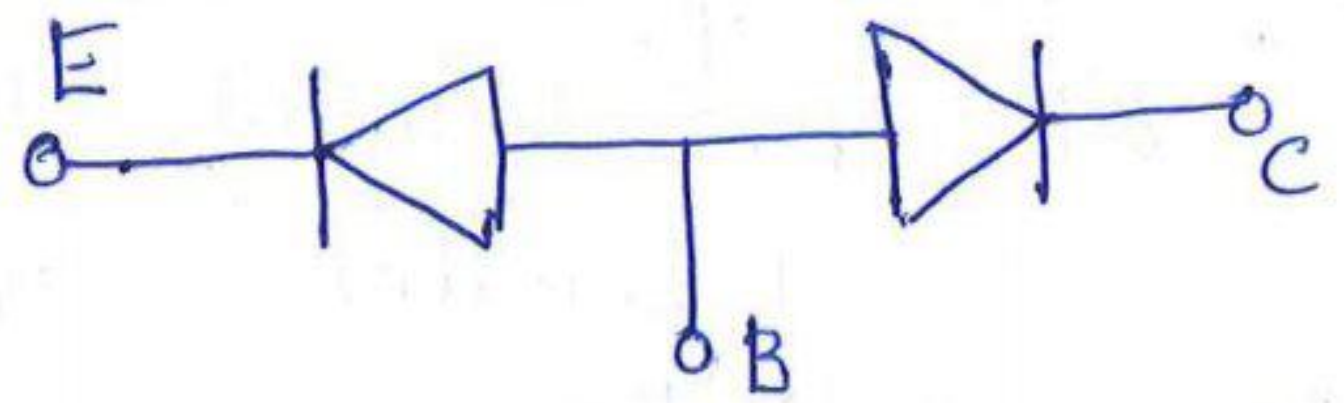


fig: Two diode transistor Analogy

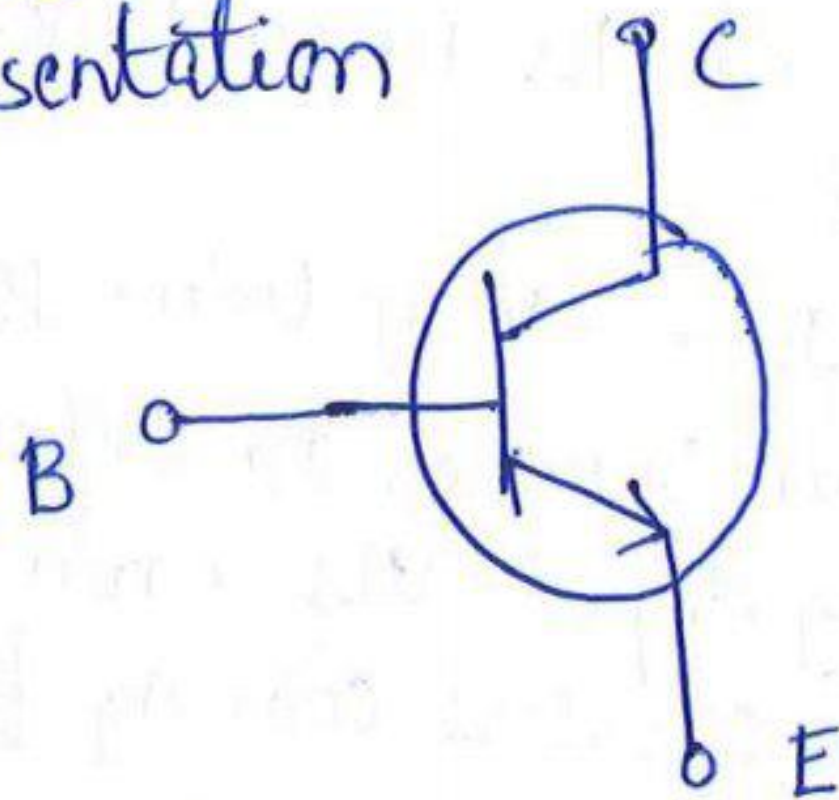


fig: Symbol of NPN Transistor

⇒ Working of NPN Transistor:

* For a transistor to work, it has to be biased by applying external power supply.

* Emitter Base junction of the transistor should be in forward bias

1. Emitter which is of N-type is connected to the -ve terminal of the battery.
2. Base which is of P-type is connected to the +ve terminal of the battery.

* Collector Base junction of the transistor should be in reverse bias. (16)

1. Collector which is of N-type is connected to the +ve terminal of the battery.
2. Base which is of P-type is connected to the -ve terminal of the battery.

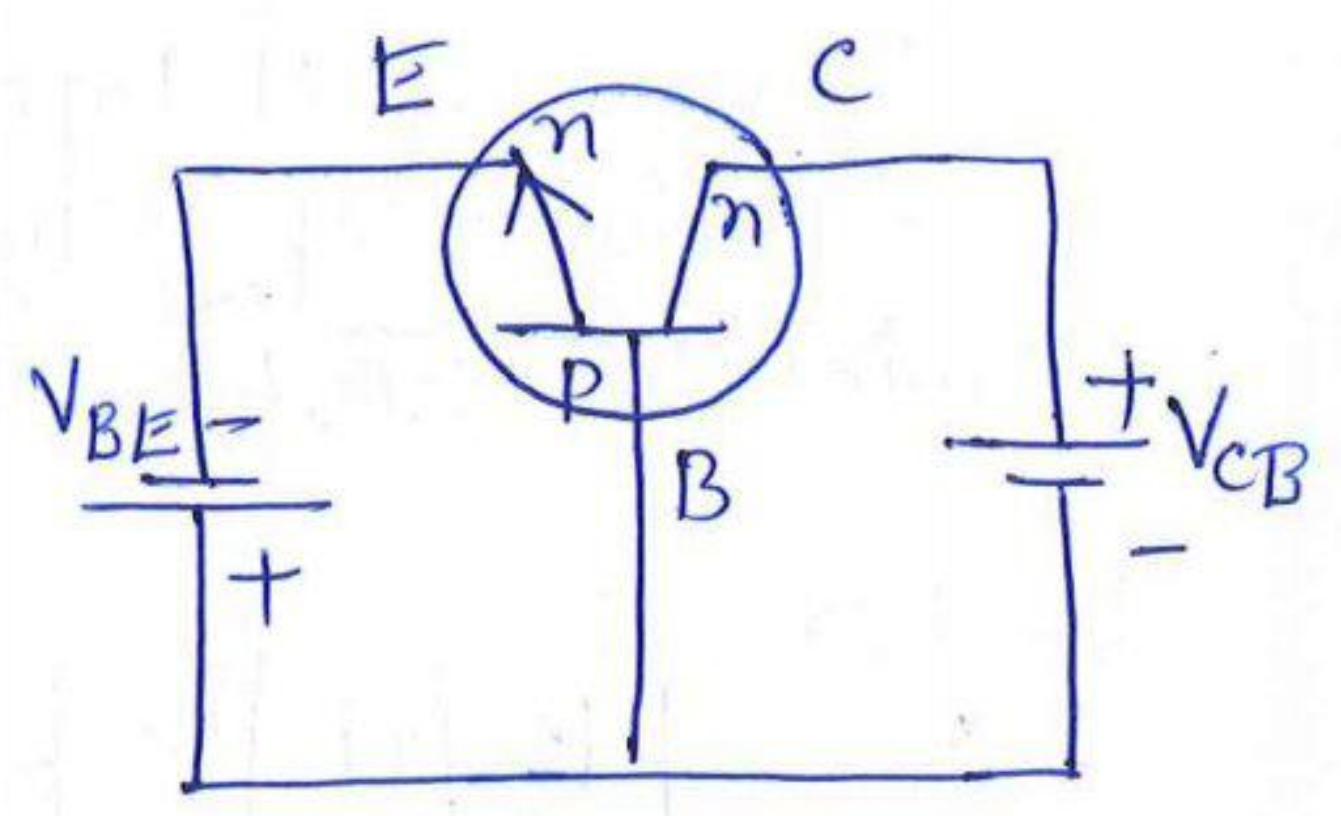
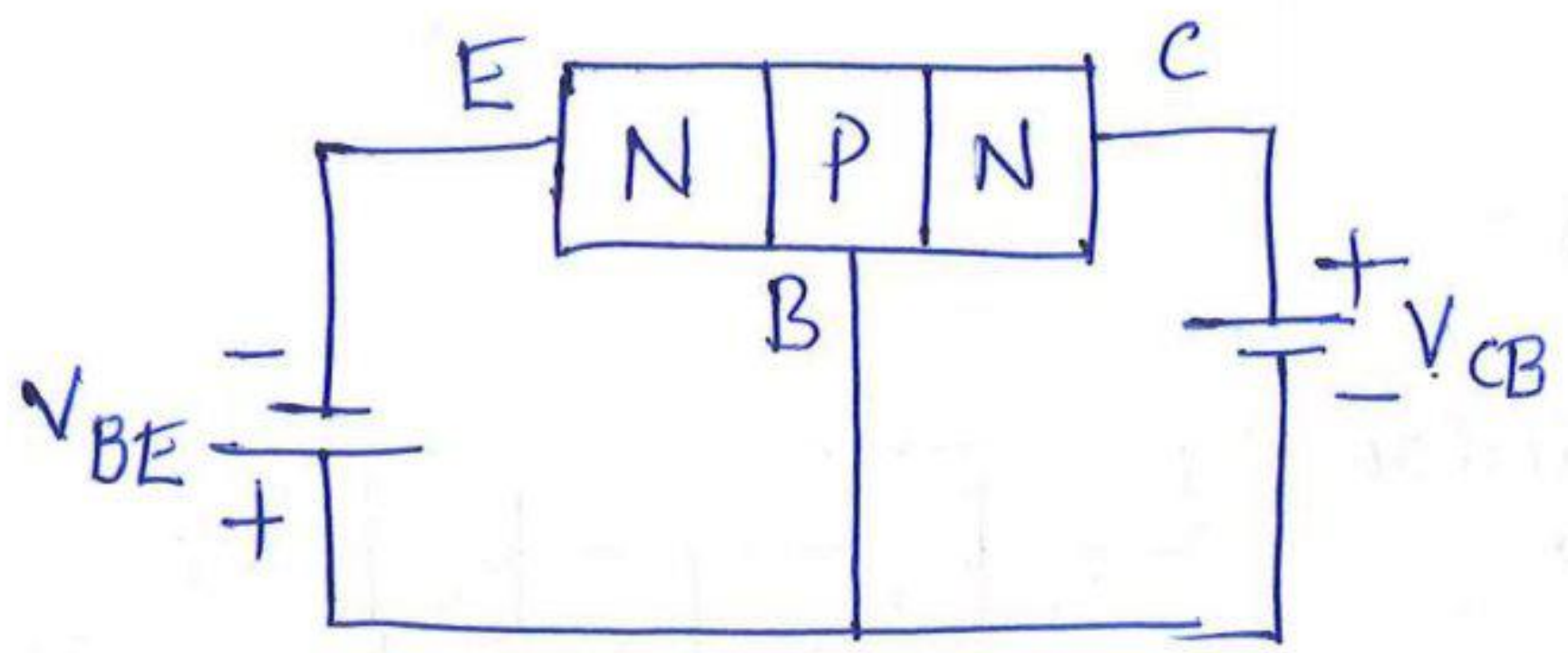
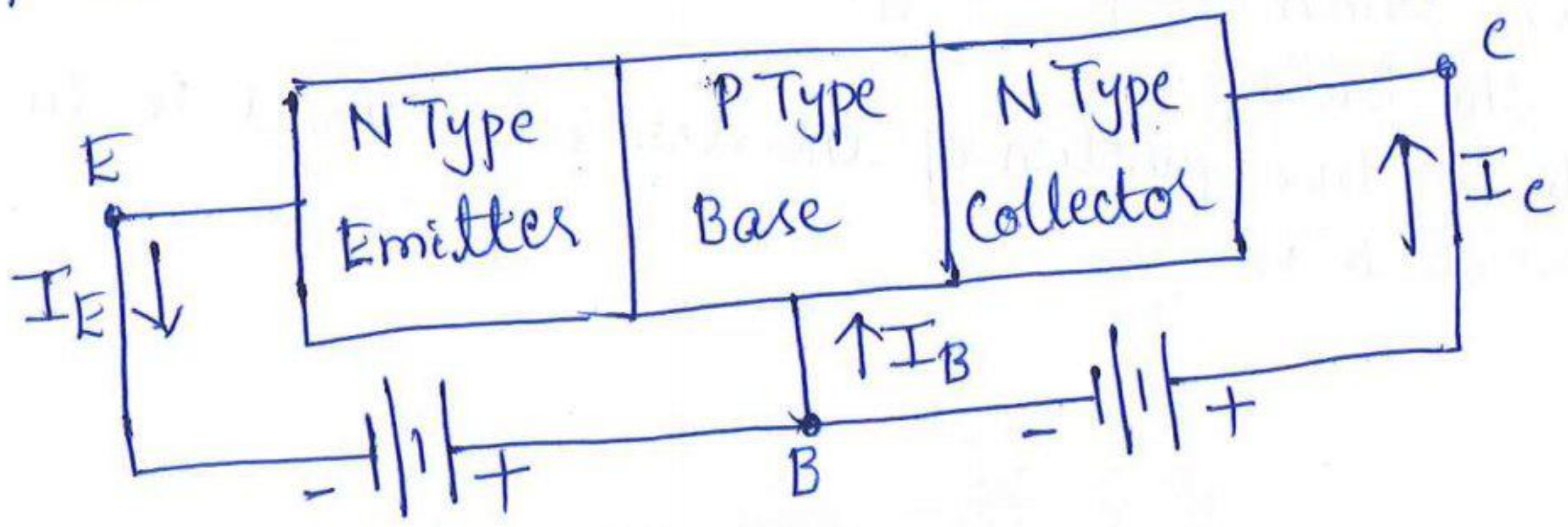


Fig: Biasing of a NPN transistor

- * Under the influence of applied voltage, electrons in the N-type Emitter get repelled from the -ve terminal of the battery.
- * Electrons flow from Emitter to the Base region.
- * As base is lightly doped, there will be small number of holes in the base.
- * Only a small percentage of electrons from the emitter will recombine with holes in the base region.
- * Remaining large number of electrons cross base region & enter into the collector region.
- * Due to large bias voltage, electrons move towards the +ve terminal of the battery & hence collector collects 98% of the electrons emitted by the emitter.



PNP TRANSISTOR

⇒ Construction of PNP Transistor

* It is formed by sandwiching a single N-region between 2 P-regions.

* Two PN junctions are connected together.

Emitter is of P-Type

Base is of N-Type

Collector is of P-Type

⇒ Symbol of PNP Transistor:

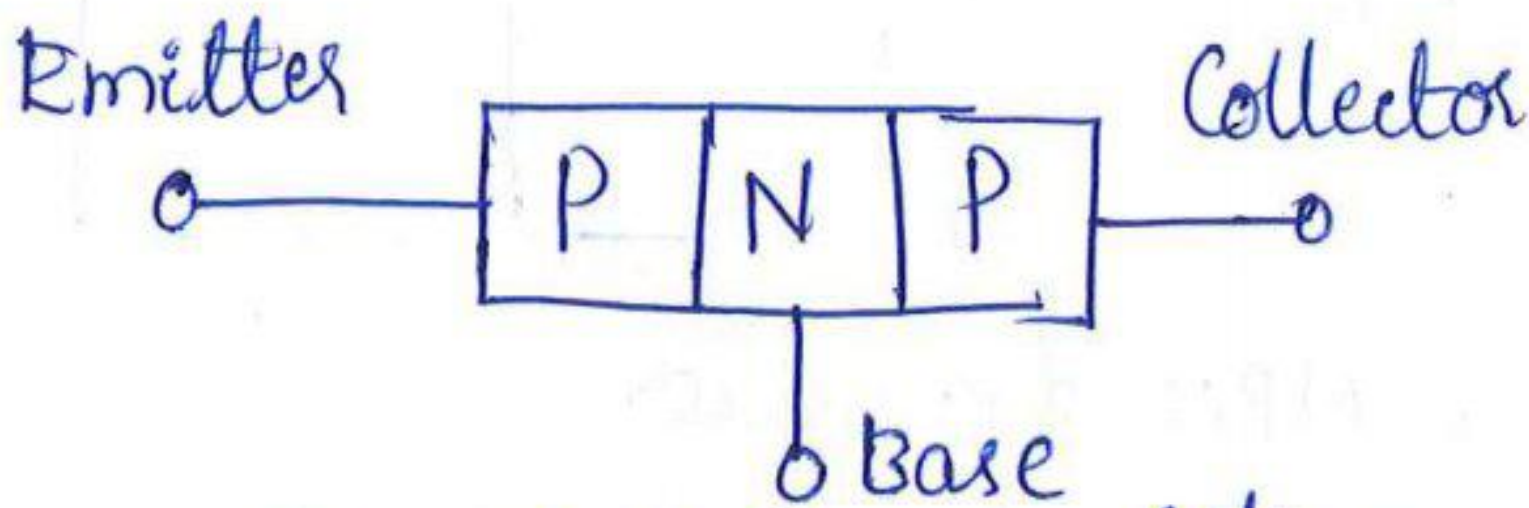


fig: Block Representation

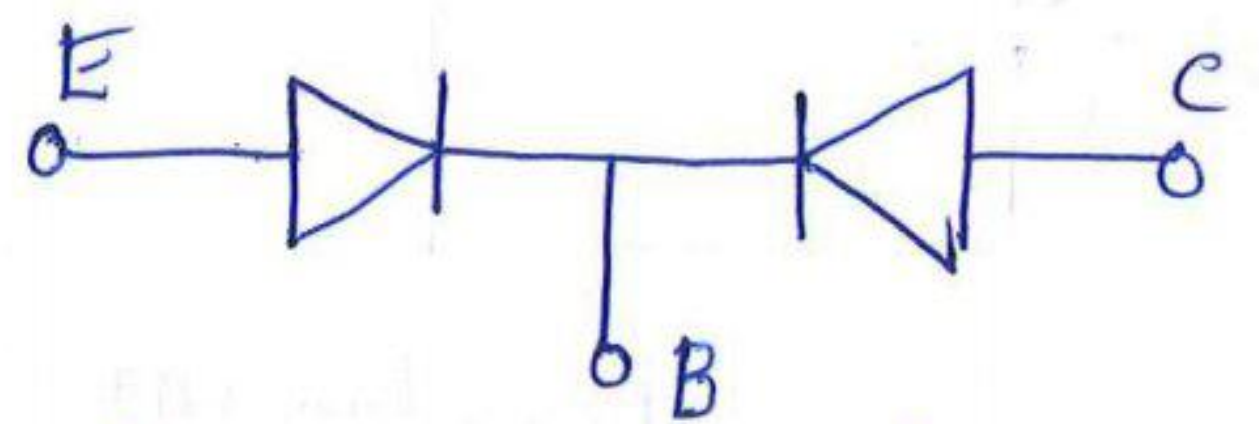


fig: Two-diode transistor analogy

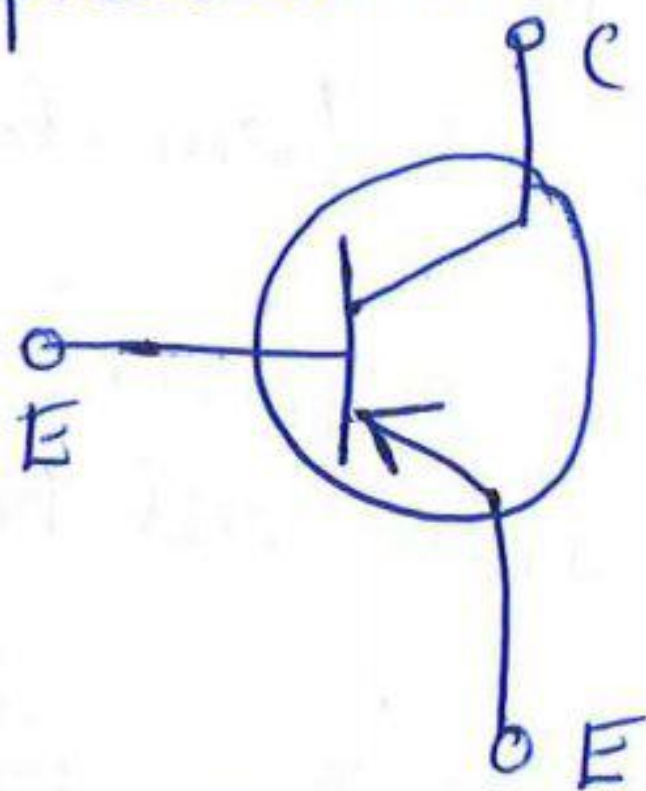


fig: Symbol of PNP Transistor

⇒ Working of PNP Transistor:

* Emitter base junction of the transistor should be in forward bias.

1. Emitter which is of P-type is connected to the +ve terminal of the battery.

2. Base which is of N-type is connected to the -ve terminal of the battery.

* Collector base junction of the transistor should be in reverse bias.

- 1. Collector which is of Ptype is connected to the -ve terminal of the battery.
- 2. Base which is of Ntype is connected to the +ve terminal of the battery.
- * Under the influence of the applied voltage, holes in the P-type Emitter get repelled from the +ve terminal of the battery.
- * Holes move from emitter to the Base region.
- * As base is very thin & lightly doped, holes from the emitter recombine with the small number of electrons in the base region.
- * Remaining large number of cross the base region & enter into the collector region as they get attracted by the -ve terminal of the battery.
- * In a NPN transistor, majority charge carriers are electrons
In a PNP transistor, majority charge carriers are holes.

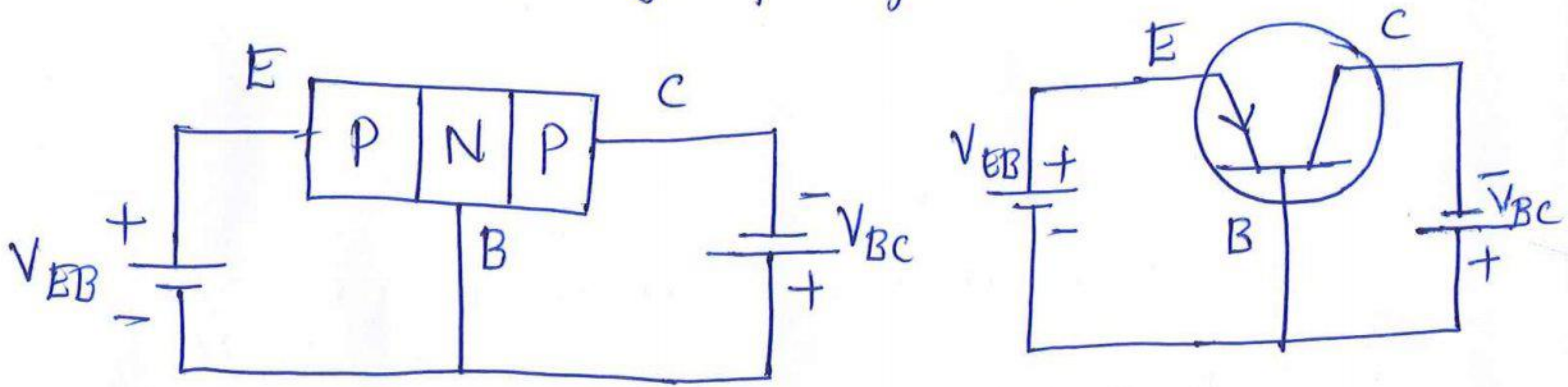
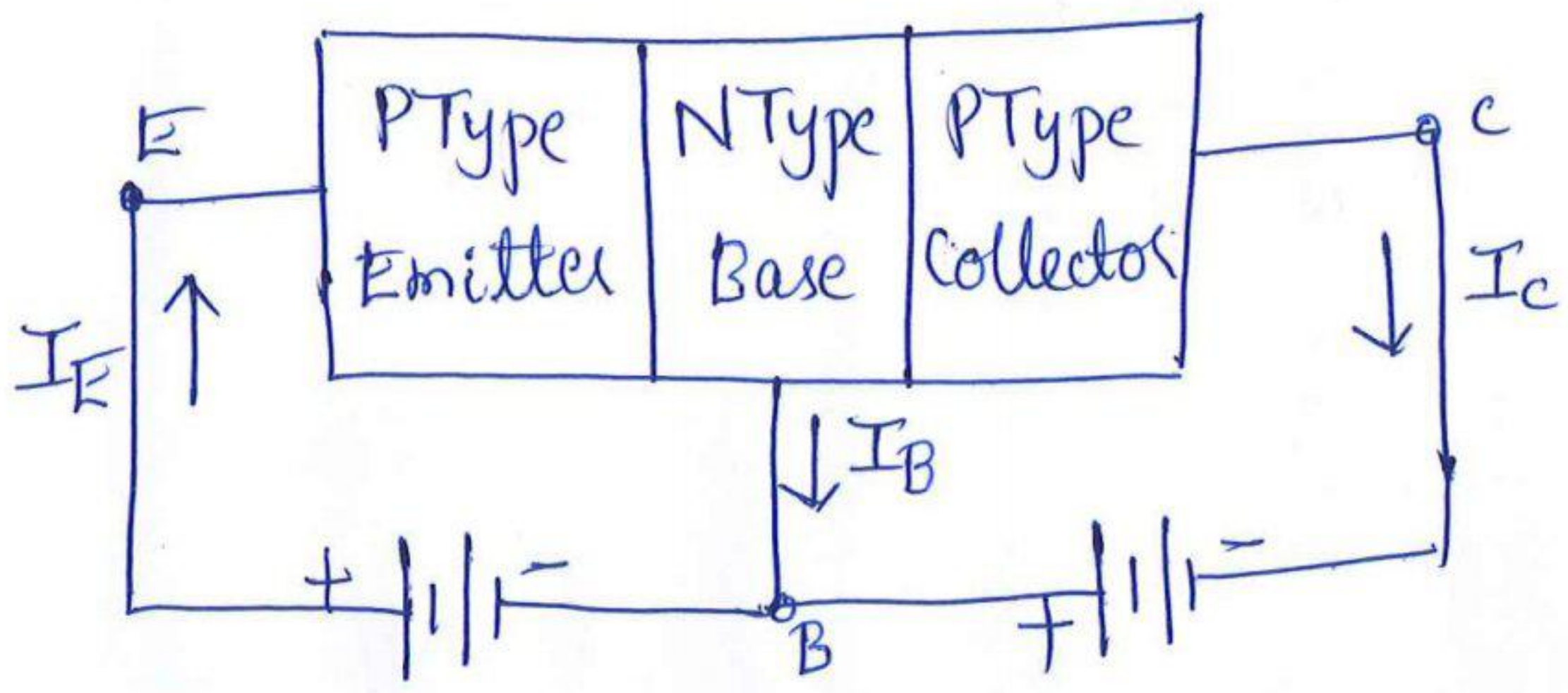


Fig: Biasing of a PNP Transistor



⇒ Transistor Current Equation:

Emitter Current = Base Current + Collector Current

$$\boxed{I_E = I_B + I_C} \quad \text{--- (1)}$$

* Current Amplification factor in CB configuration

$$\boxed{\alpha = \frac{\Delta I_C}{\Delta I_E}} \quad \text{--- (2)}$$

It is defined as the ratio of change in o/p current to the change in the i/p current.

* Current Amplification factor in CE configuration

$$\boxed{\beta = \frac{\Delta I_C}{\Delta I_B}} \quad \text{--- (3)}$$

* Current Amplification factor in CC configuration

$$\boxed{\gamma = \frac{\Delta I_E}{\Delta I_B}} \quad \text{--- (4)}$$

* We know $I_E = I_B + I_C$

$$\Delta I_E = \Delta I_B + \Delta I_C$$

$$\Delta I_B = \Delta I_E - \Delta I_C \quad \text{--- (5)}$$

substituting the value of ΔI_B in β (i.e., substitute equ(5) in equ(3))

$$\beta = \frac{\Delta I_C}{\Delta I_E - \Delta I_C} \quad \text{Divide both numerator \& denominator with } \Delta I_E$$

$$\beta = \frac{\Delta I_C / \Delta I_E}{\Delta I_E / \Delta I_E - \frac{\Delta I_C}{\Delta I_E}} = \frac{\frac{\Delta I_C}{\Delta I_E}}{1 - \frac{\Delta I_C}{\Delta I_E}} \quad \text{since } \alpha = \frac{\Delta I_C}{\Delta I_E}$$

$$\boxed{\beta = \frac{\alpha}{1 - \alpha}}$$

* Adding '1' on both sides

$$\beta + 1 = \frac{\alpha}{1 - \alpha} + 1 = \frac{\alpha + 1 - \alpha}{1 - \alpha}$$

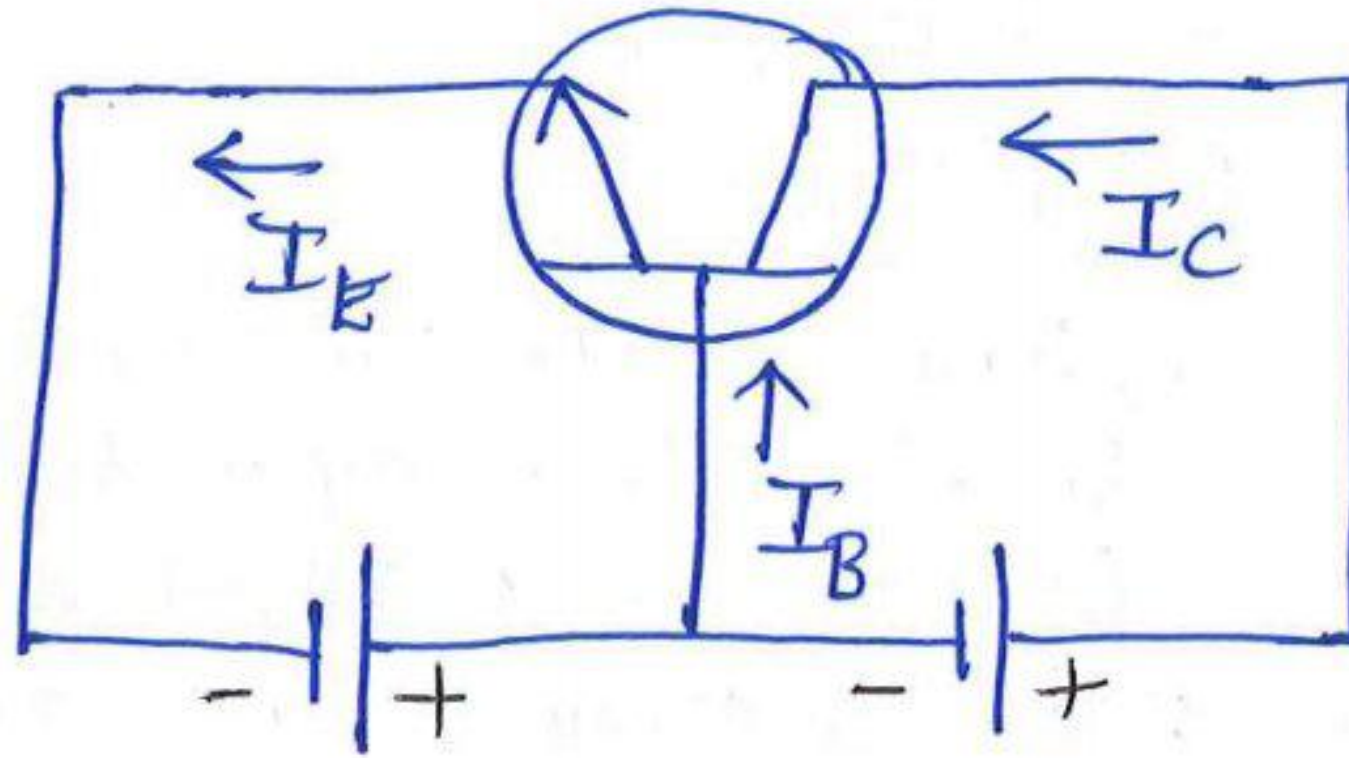
$$\beta + 1 = \frac{1}{1 - \alpha}$$

$$1 - \alpha = \frac{1}{1 + \beta}$$

$$\alpha = 1 - \frac{1}{1 + \beta} = \frac{1 + \beta - 1}{1 + \beta}$$

$$\Rightarrow \boxed{\alpha = \frac{\beta}{1 + \beta}}$$

Problem: A n-p-n transistor is provided with biasing voltage V_{BE} and V_{CB} . Calculate the values of I_C and I_E if α_{dc} is 0.96 & I_B is $80 \mu A$. Calculate the value of β_{dc} .



Solution:

$$\alpha_{dc} = 0.96, I_B = 80 \times 10^{-6} A$$

$$\beta_{dc} = \frac{\alpha_{dc}}{1 - \alpha_{dc}} = \frac{0.96}{1 - 0.96} = \frac{0.96}{0.04} = 24$$

We know $\beta = \frac{I_C}{I_B} \Rightarrow I_C = \beta I_B$

$$I_C = \frac{\alpha}{1 - \alpha} I_B$$

$$I_C = \frac{0.96}{1 - 0.96} \times 80 \times 10^{-6} = 1.92 \text{ mA}$$

We know $\alpha_{dc} = \frac{I_C}{I_E} \Rightarrow I_E = \frac{I_C}{\alpha_{dc}}$

$$I_E = \frac{1.92 \text{ mA}}{0.96}$$

$$I_E = 2 \text{ mA}$$

⇒ Transistor Configurations:

Transistor can be connected in the circuit in 3 different ways. By making one among the three terminals as a common point gives rise to 3 transistor configurations.

1. Common Base Configuration

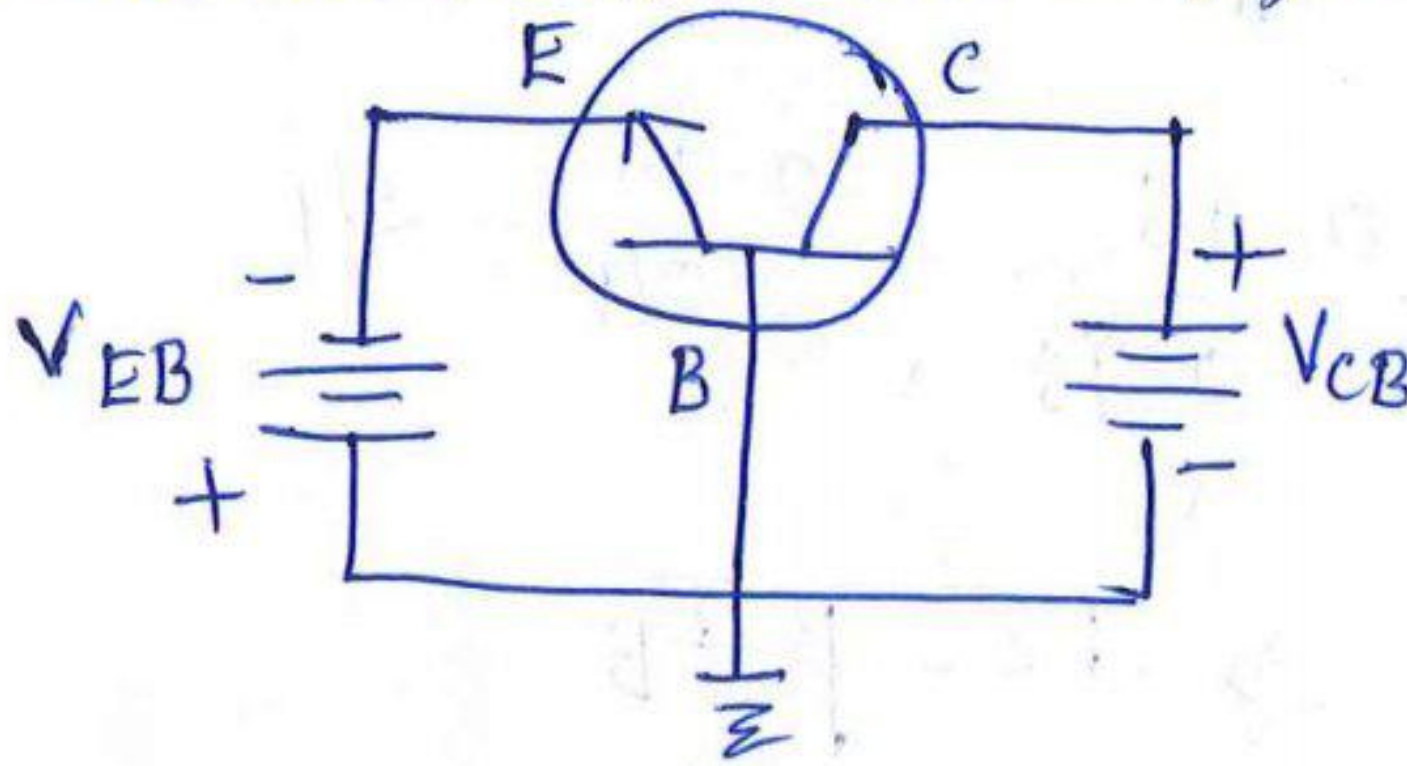
2. Common Emitter Configuration

3. Common Collector Configuration

⇒ Common Base Configuration:

* In this configuration, Base acts as common terminal
Emitter acts as input terminal
Collector acts as output terminal.

* As common terminal is always connected to ground, the base terminal is connected to ground, hence it is called Common Base (or) Grounded Base configuration



* Input voltage is applied between base & emitter terminals

* Output is taken from base & collector terminals.

* Voltage gain A_v of the transistor is defined as the ratio of o/p voltage to i/p voltage.

* Current Amplification factor is defined as the ratio of o/p current to i/p current

$$\alpha = \frac{I_C}{I_E}$$

(or)

$$\alpha = \frac{\Delta I_C}{\Delta I_E}$$

* Input characteristics is taken between i/p current I_E & i/p voltage V_{EB} by making V_{CB} (collector base voltage) as constant

* Output characteristics is taken between o/p voltage V_{CB} & i/p current I_C by making I_E (input current) as constant

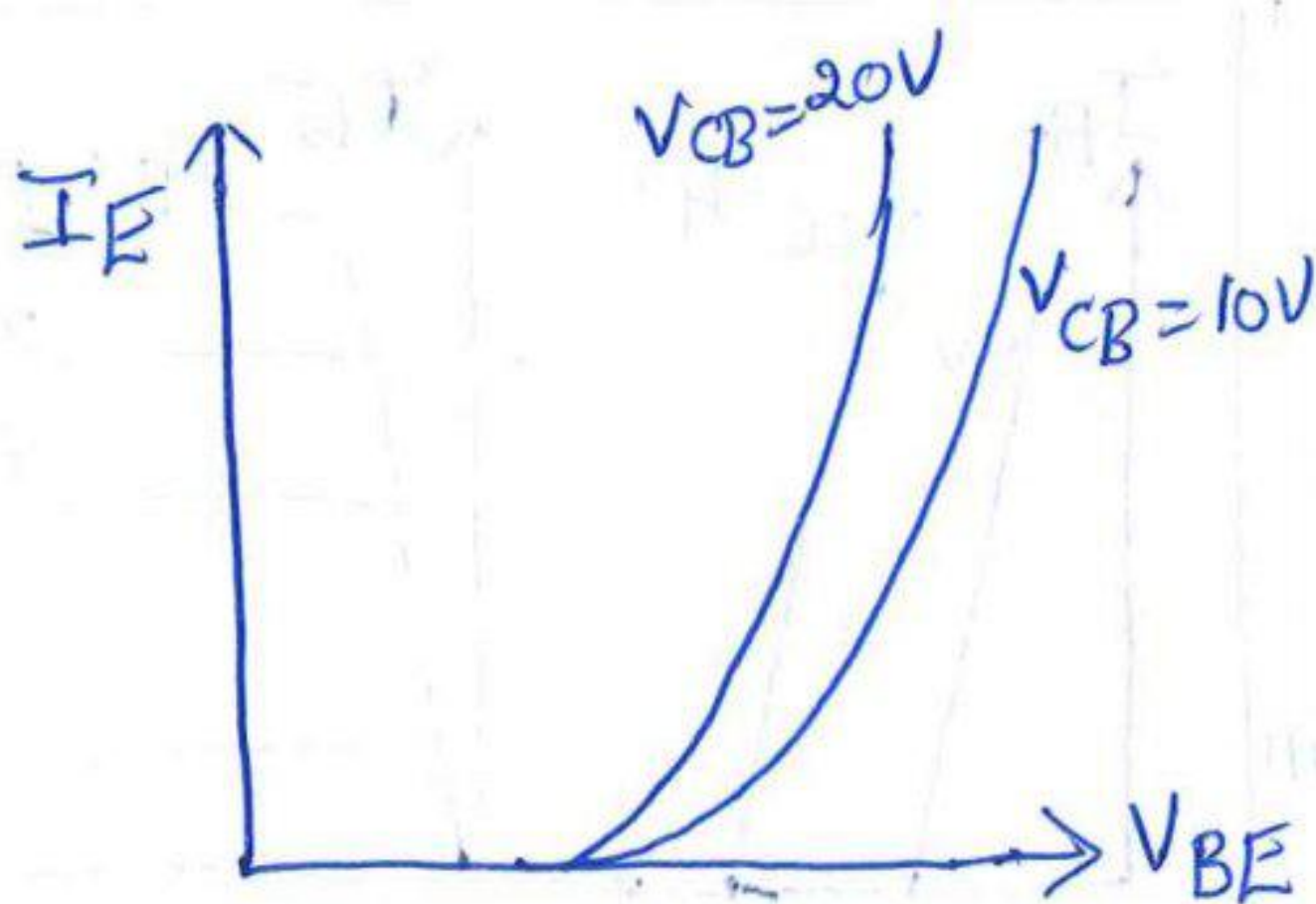


fig: Input characteristics

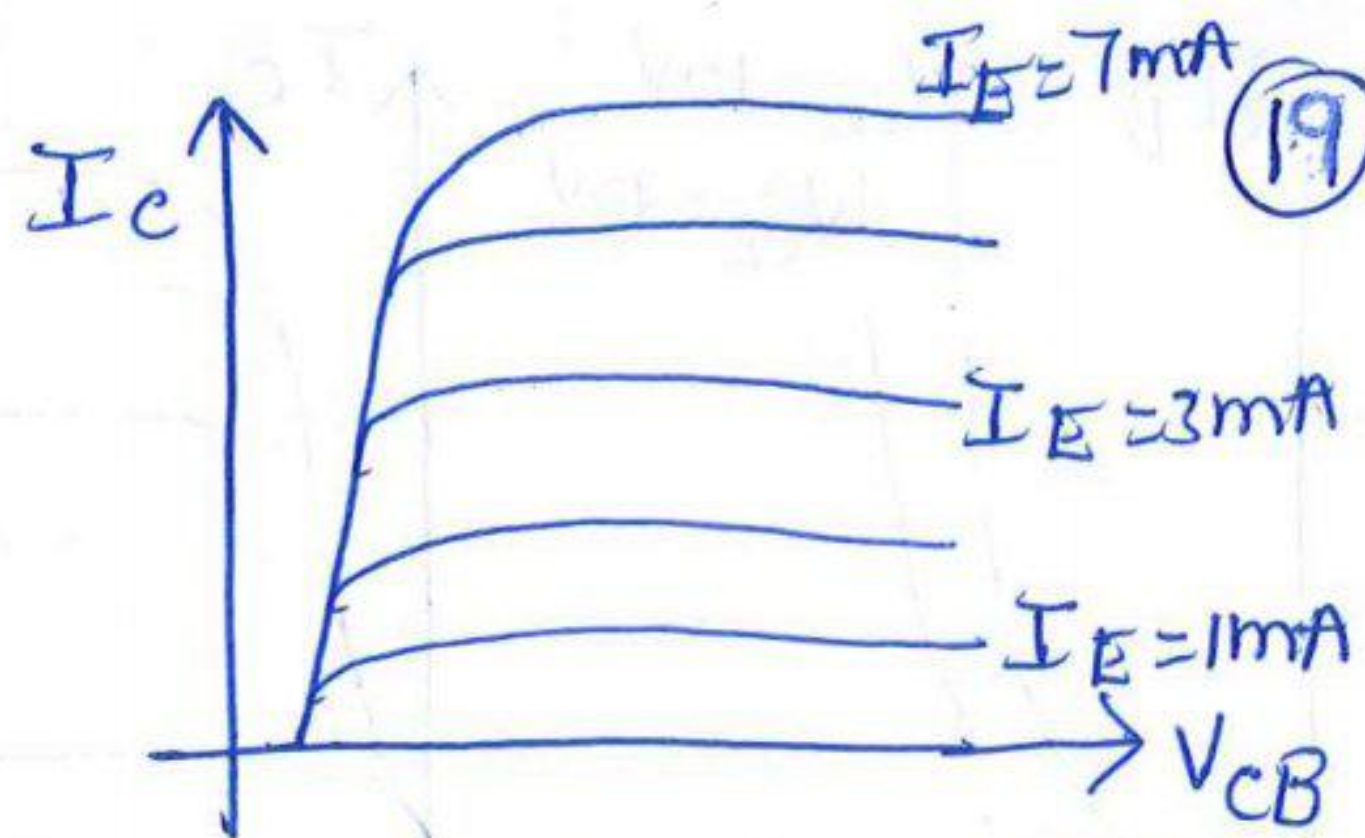
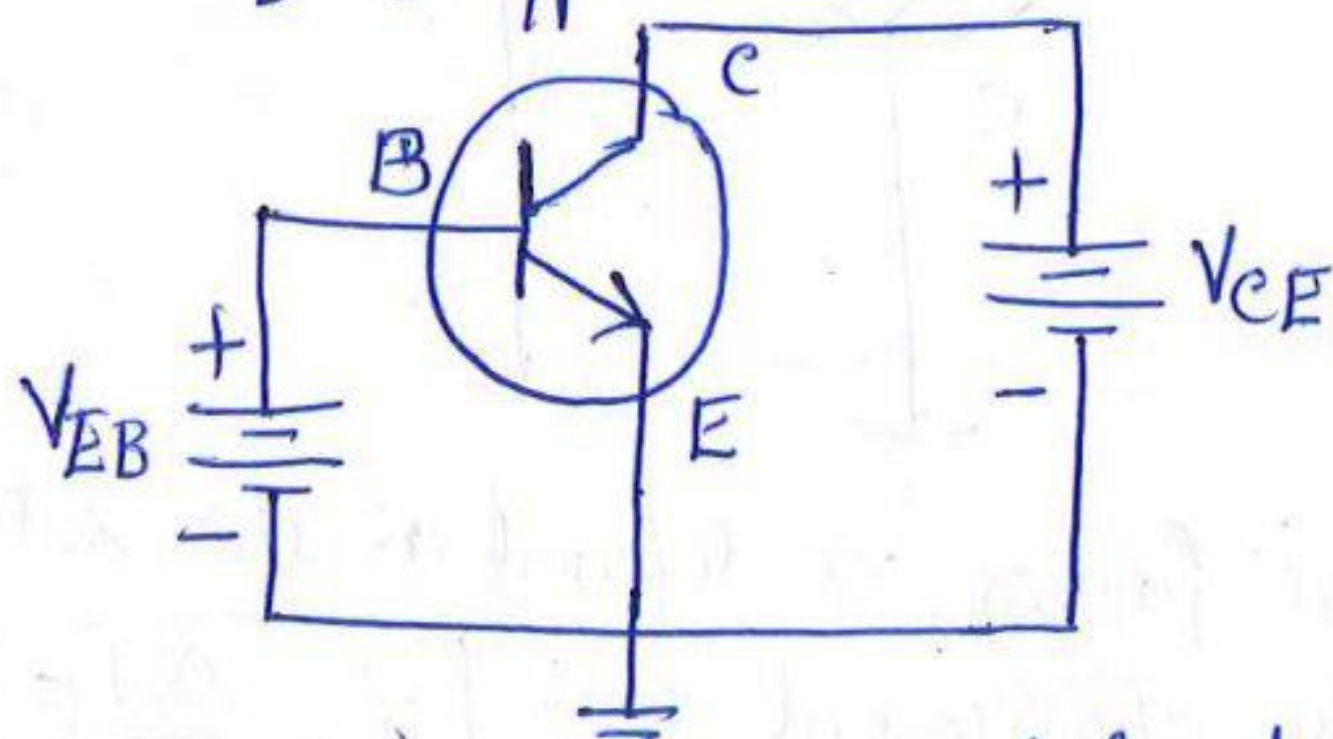


fig: Output characteristics

⇒ Common Emitter Configuration:

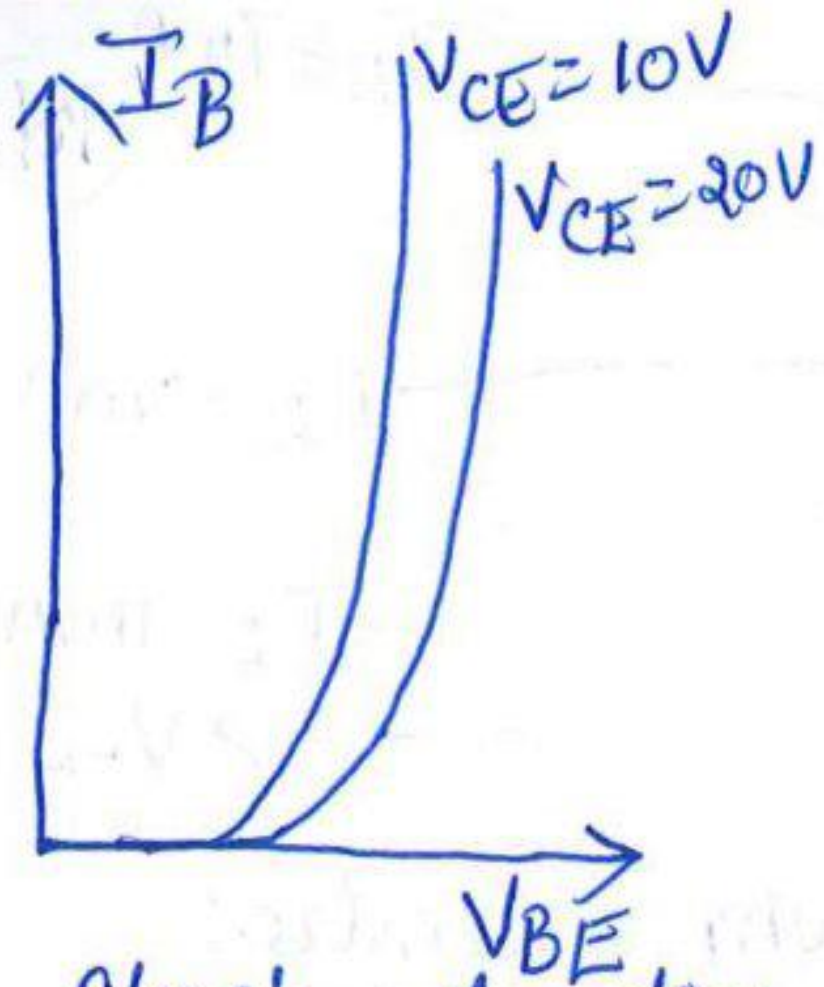
- * It is also known as grounded emitter configuration.
- * In this configuration, Emitter acts as a Common terminal
Base acts as input terminal
Collector acts as output terminal
- * As emitter is connected to ground, it is called grounded emitter configuration.
- * As emitter is common to both input and output circuits, it is called common emitter configuration.
- * I/p voltage is Base Emitter Voltage V_{EB}
O/p voltage is collector Emitter Voltage V_{CE}
- * I/p current is I_B & o/p current is I_C



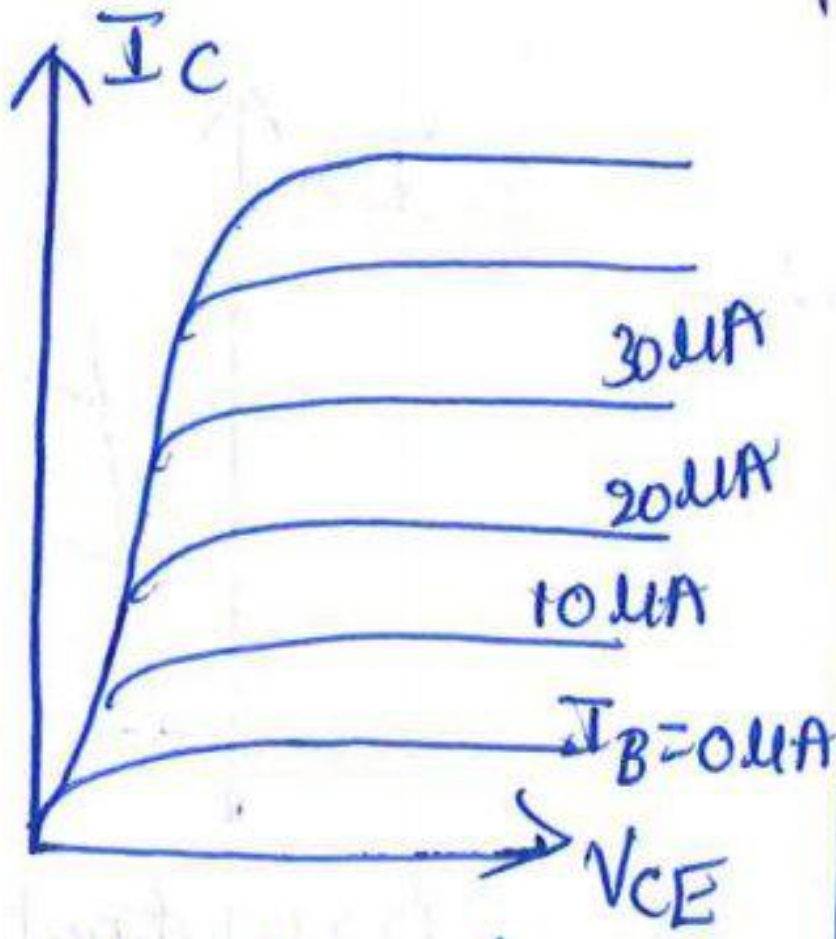
* Current Amplification factor is defined as the ratio of change in o/p current to the change in i/p current

$$\beta = \frac{\Delta I_C}{\Delta I_B}$$

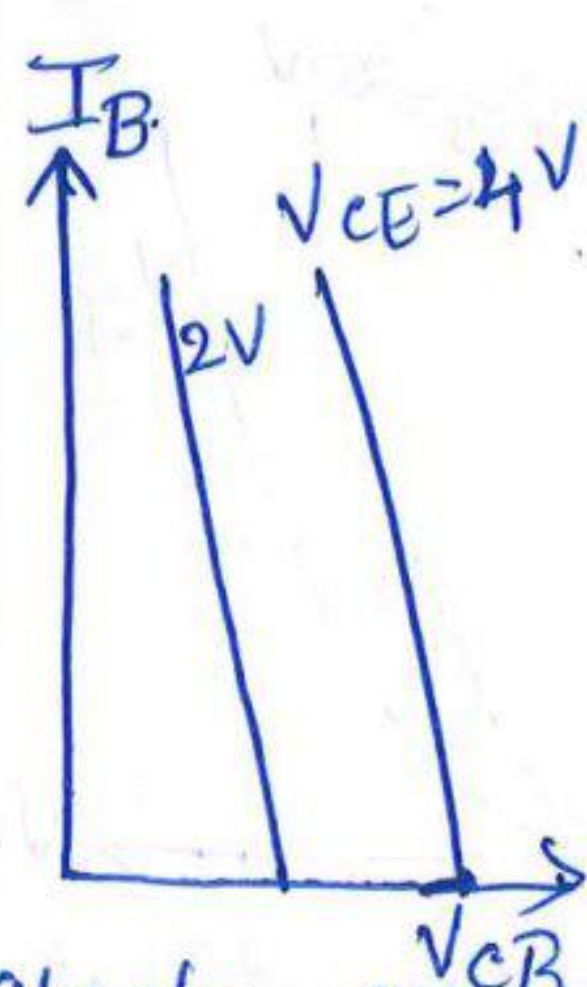
- * Input characteristics is taken between i/p current I_B & i/p voltage V_{BE} by making o/p voltage V_{CE} as constant.
- * Output characteristics is taken between o/p voltage V_{CE} & o/p current I_C by making i/p current I_B as constant.



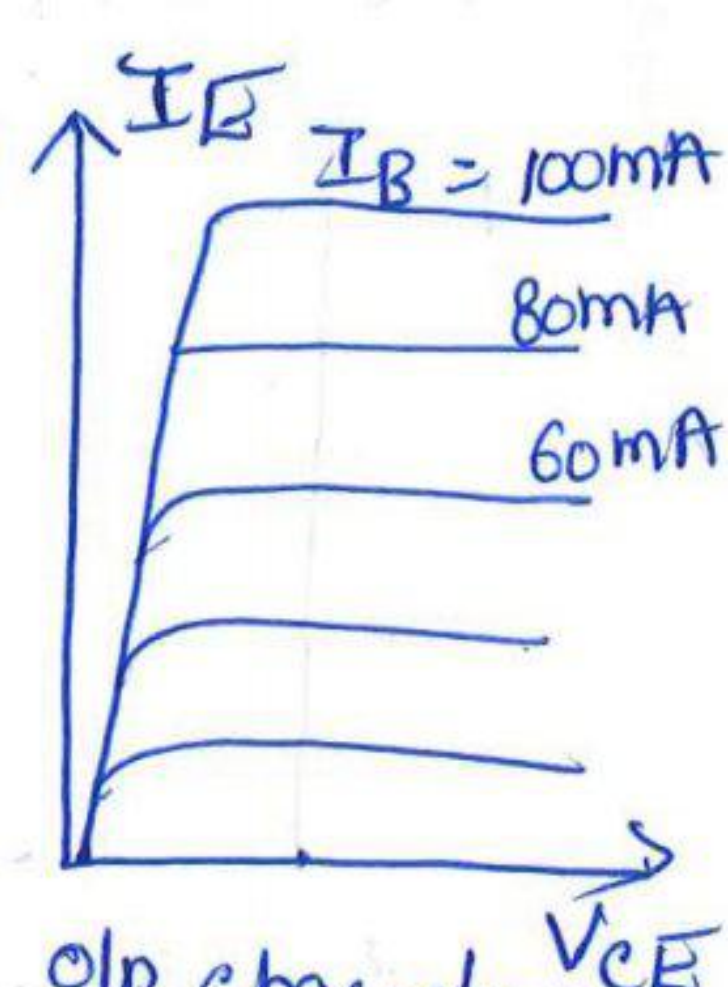
i/p characteristics of CE



o/p characteristics of CE



i/p characteristics of CC



o/p characteristics of CC

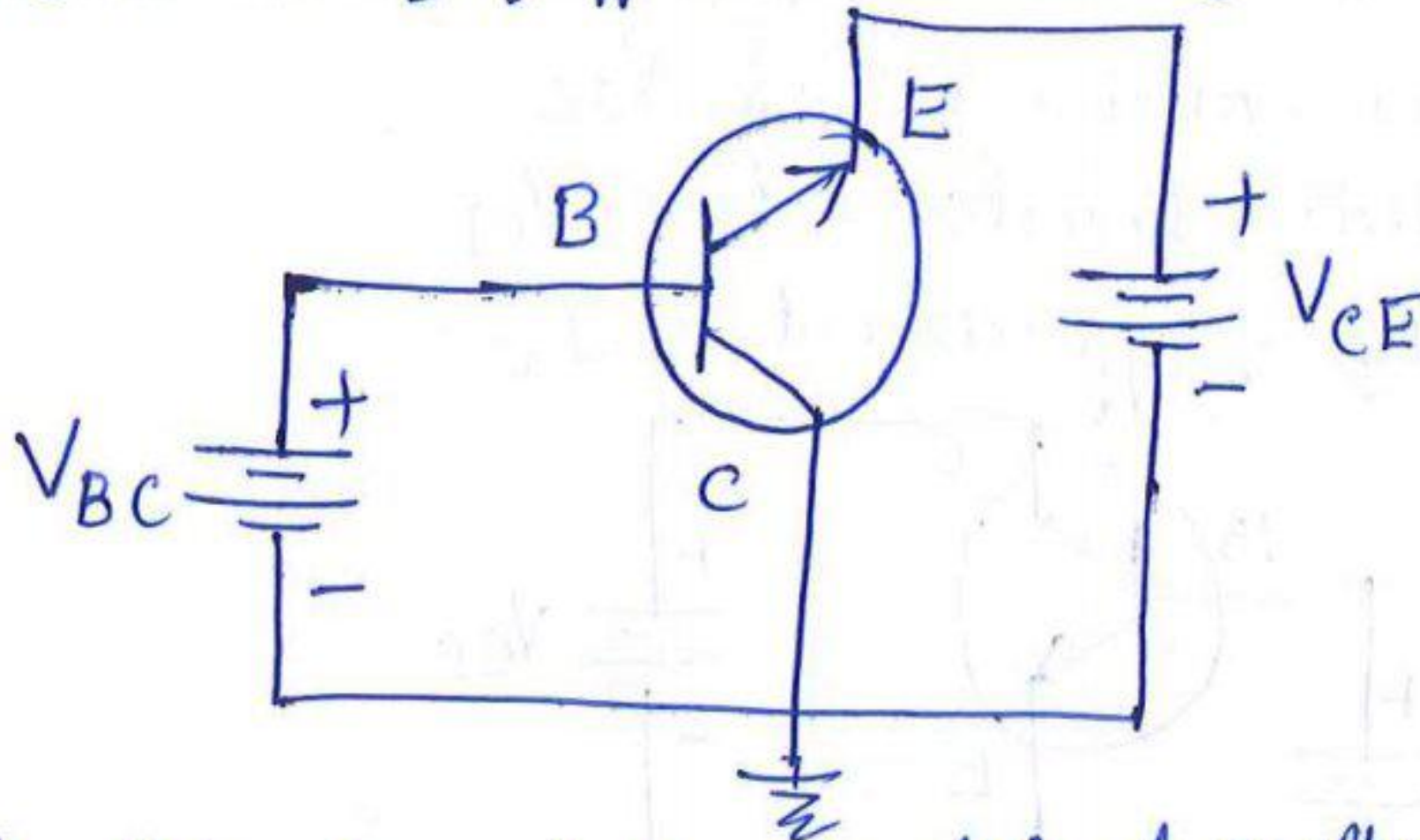
⇒ Common Collector Configuration:

* As the collector terminal is grounded, it is also called as grounded collector configuration.

* In this configuration, Collector acts as common terminal
Base acts as input terminal
Emitter acts as output terminal

* Input voltage is base collector voltage V_{BC}
Output voltage is emitter collector voltage V_{CE} .

* Input current is I_B & o/p current is I_E .



* Current Amplification factor is defined as the ratio of o/p current I_E to the i/p current I_B .

$$\beta = \frac{\Delta I_E}{\Delta I_B}$$

* Input characteristics are taken between i/p voltage V_{BC} and i/p current I_B by making o/p voltage V_{CE} as constant.

* Output characteristics are taken between o/p voltage V_{CE} & o/p current I_E by making i/p current I_B as constant.

Field Effect Transistor

⇒ Definition: FET is a unipolar device (current conduction takes place due to only majority carriers)

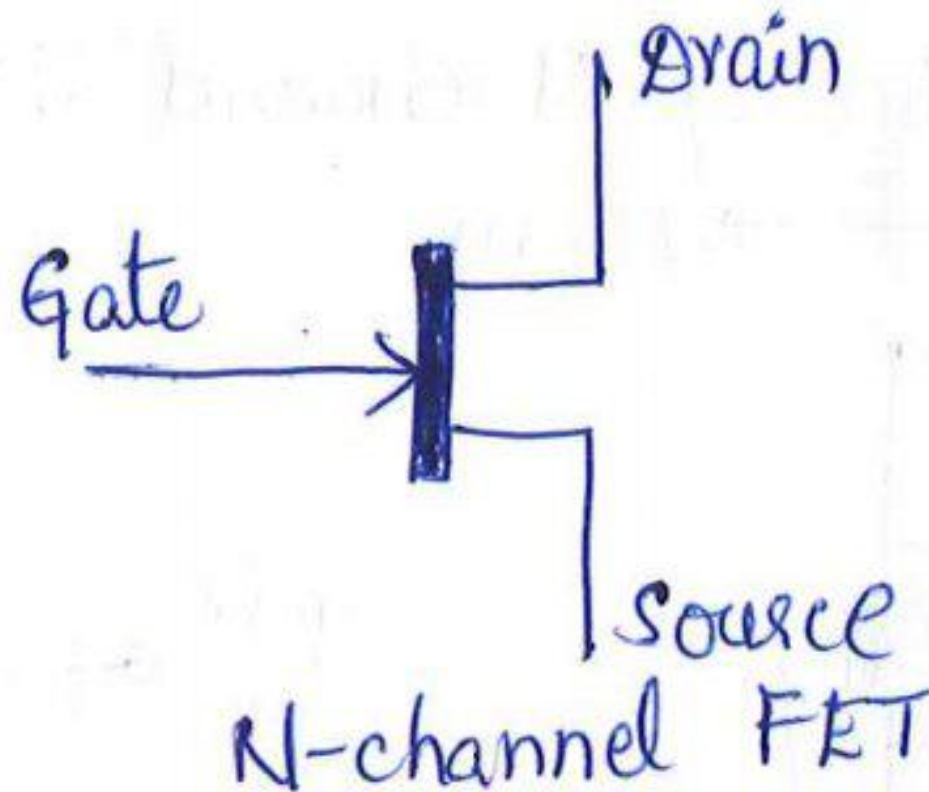
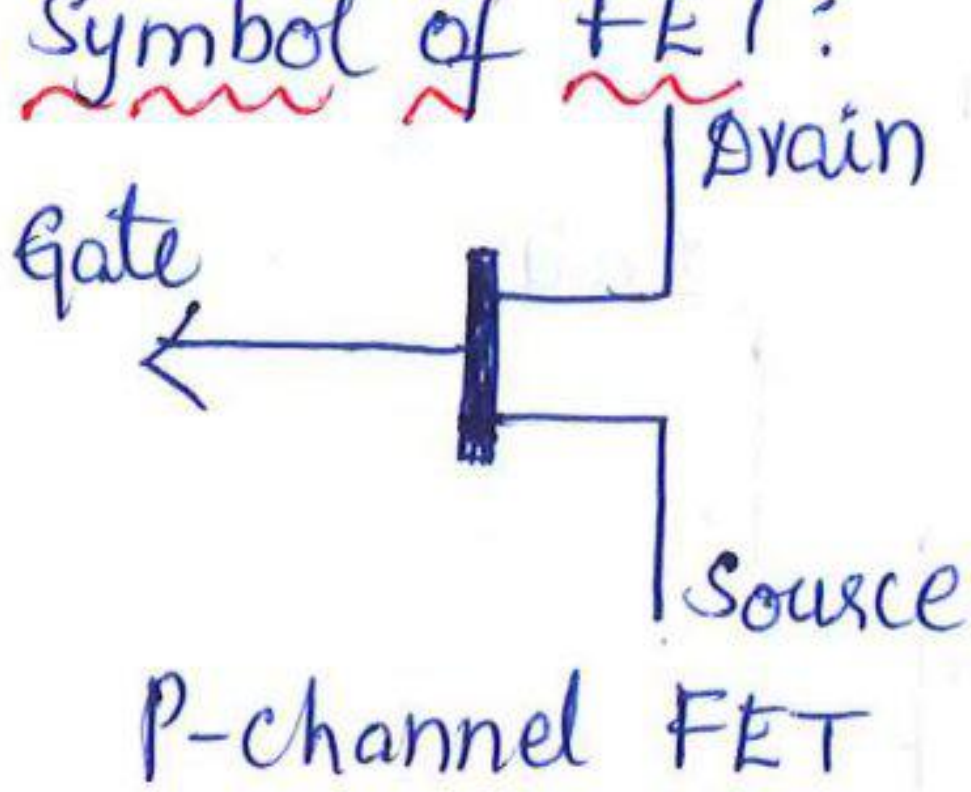
* It is a voltage controlled device.

* The name field effect is derived from the fact that the o/p current flow is controlled by an electric field setup in the device by an externally applied voltage.

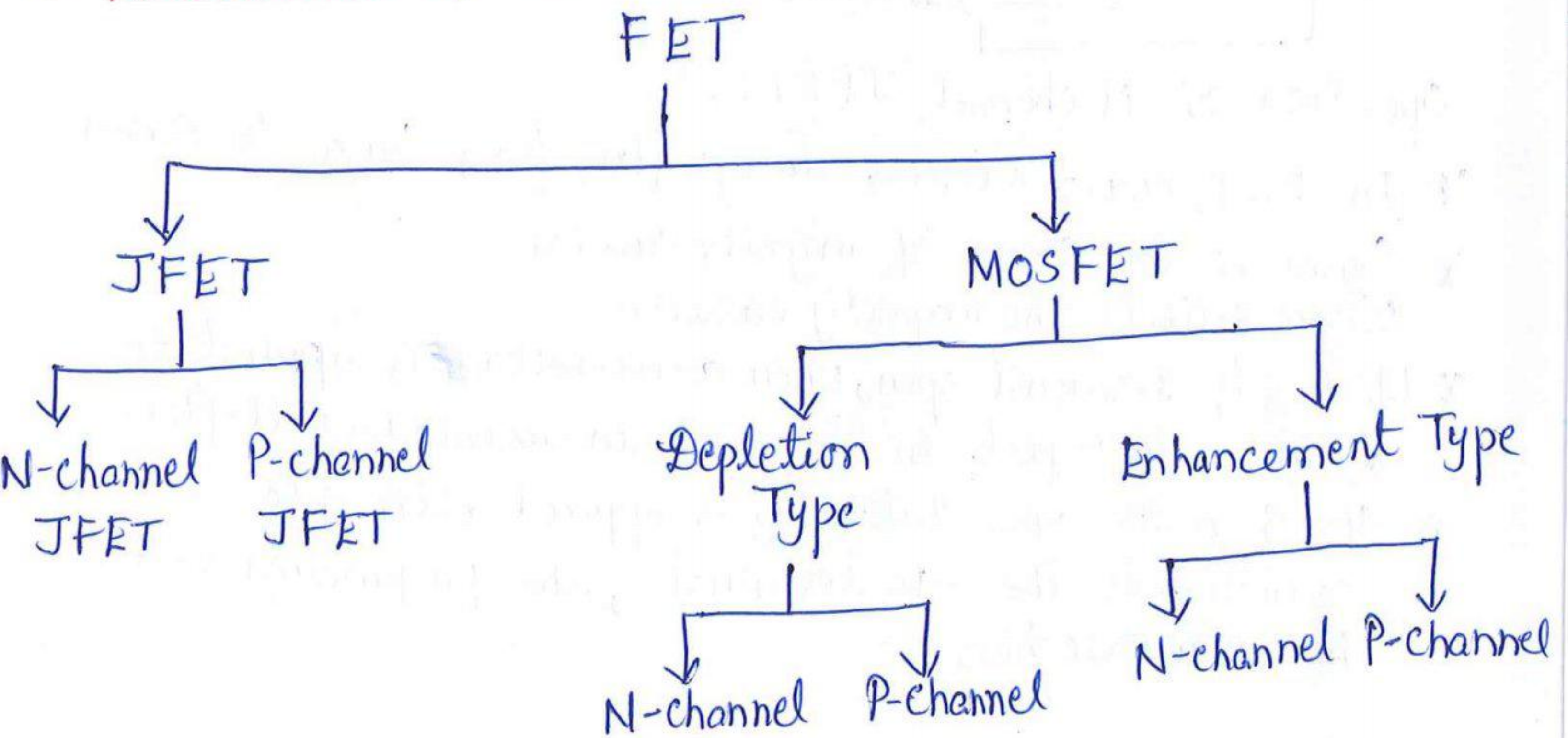
⇒ Terminals of FET: The three terminals of FET are

- 1. Gate
- 2. Source
- 3. Drain

⇒ Symbol of FET:



⇒ Classification of FET:



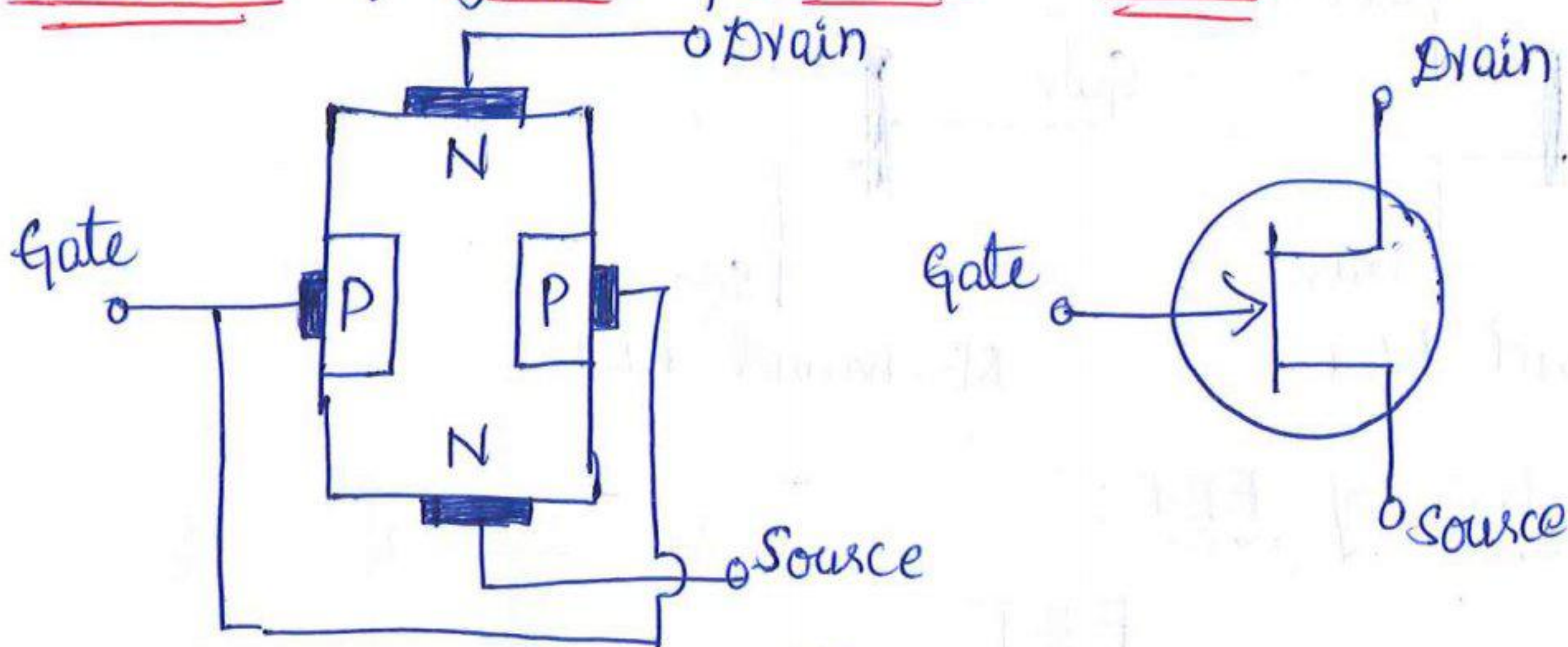
⇒ JFET (Junction Field Effect Transistor)

- * There are two types of JFET. 1. N-channel JFET
2. P-channel JFET

⇒ N-channel JFET: Construction:

- * This is made up of a N-type semiconductor material called channel which is sandwiched between two P-type materials.
- * The two P-type materials are connected together to form a gate terminal.
- * The N-type material has two terminals. The top terminal is called drain. The bottom terminal is called source.

Structure & Symbol of N-channel JFET:



Operation of N-channel JFET:

- * In FET, charge carriers always flow from Source to Drain.
- * Source is the source of majority carriers.
Drain collects the majority carriers.
- * With gate terminal open, when a +ve voltage is applied to drain with respect to source, drain current I_D will flow.
- * If Gate to Source Voltage V_{GS} is applied with Gate connected to the -ve terminal, the pn junction will be in reverse bias.

- * Due to reverse bias, depletion region will expand & penetrate into the channel from both the sides.
 - * If V_{GS} is increased, penetration will be so high such that it stops the flow of current I_D through the channel.
 - * Hence by varying V_{GS} voltage, current I_D can be varied.
 - * It can be concluded that voltage (V_{GS}) controls the current flowing through the channel.
- Hence FET is called Voltage Controlled device.

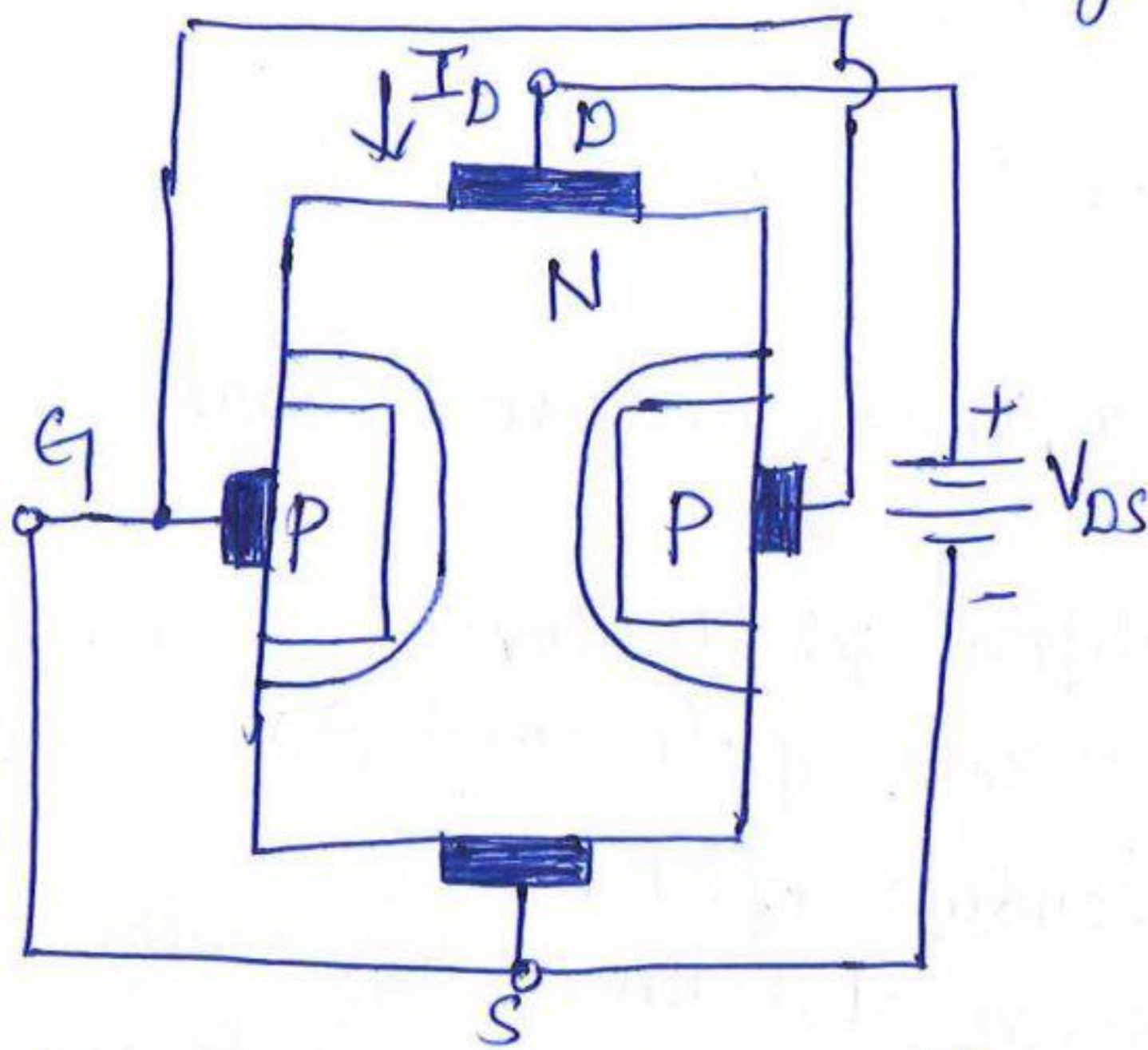


fig: A JFET with no gate to source voltage

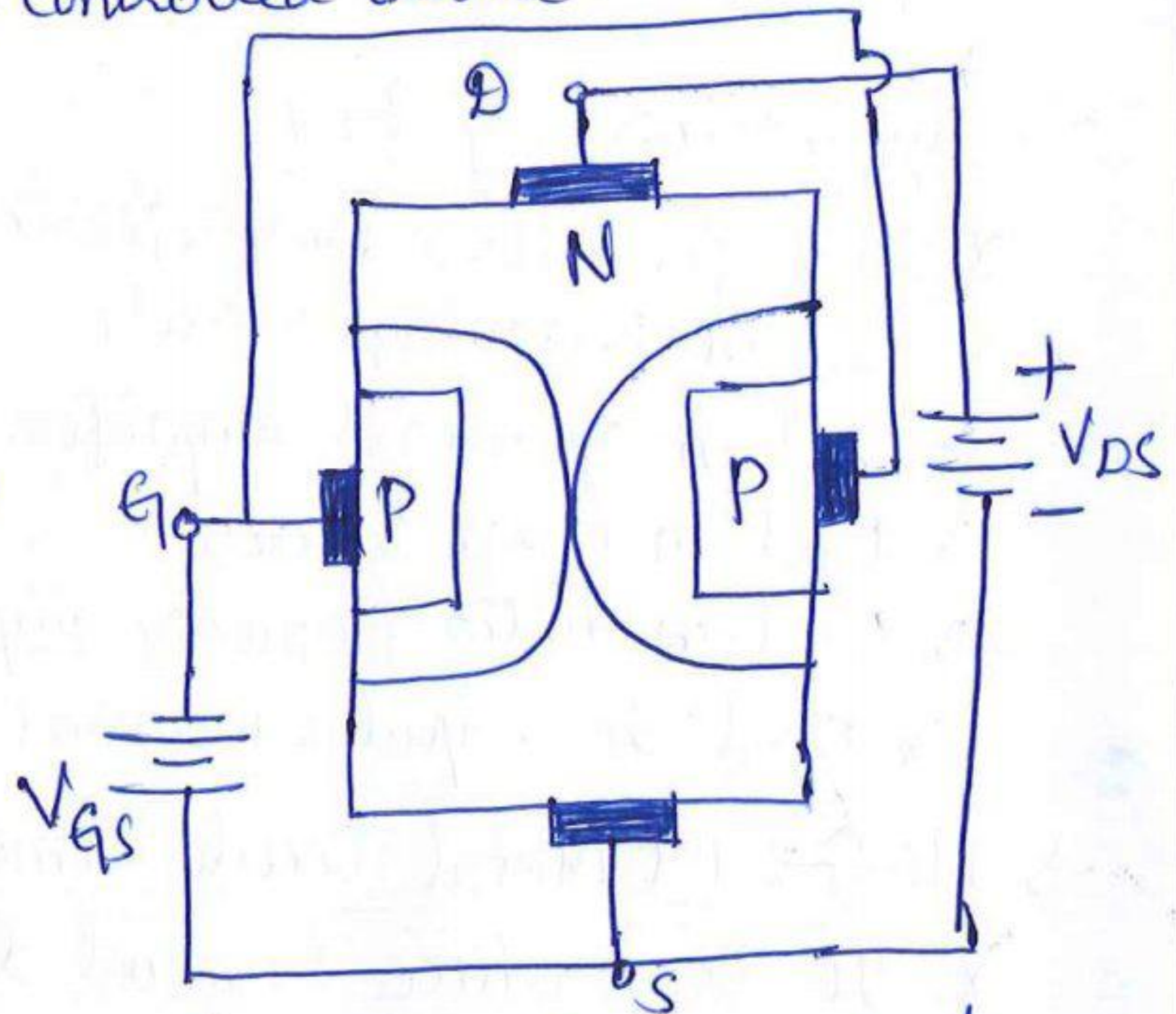


fig: When V_{GS} is applied

* As the channel is made up of N-type material, the device is called N-channel JFET.

Characteristics of JFET:

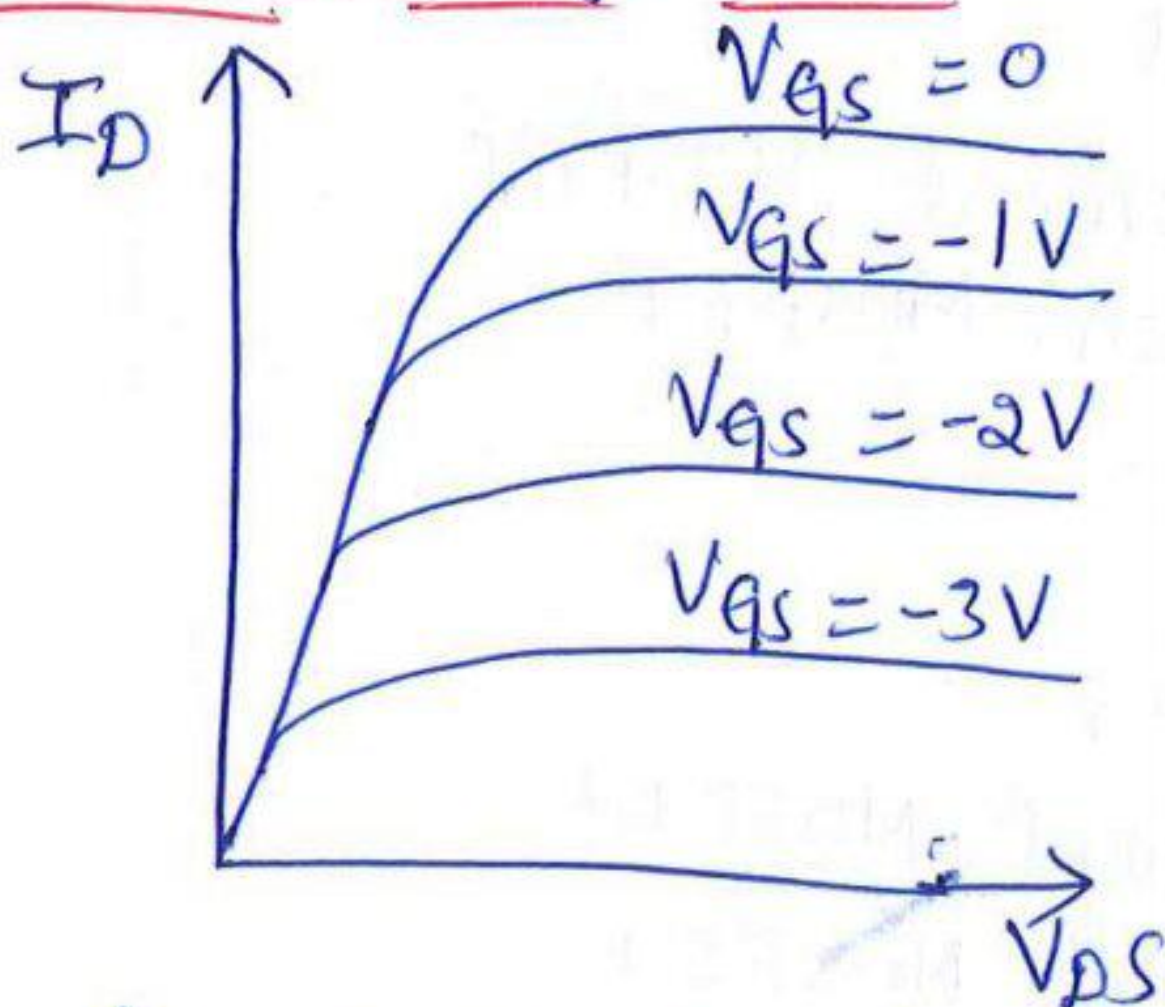


fig: Drain characteristics

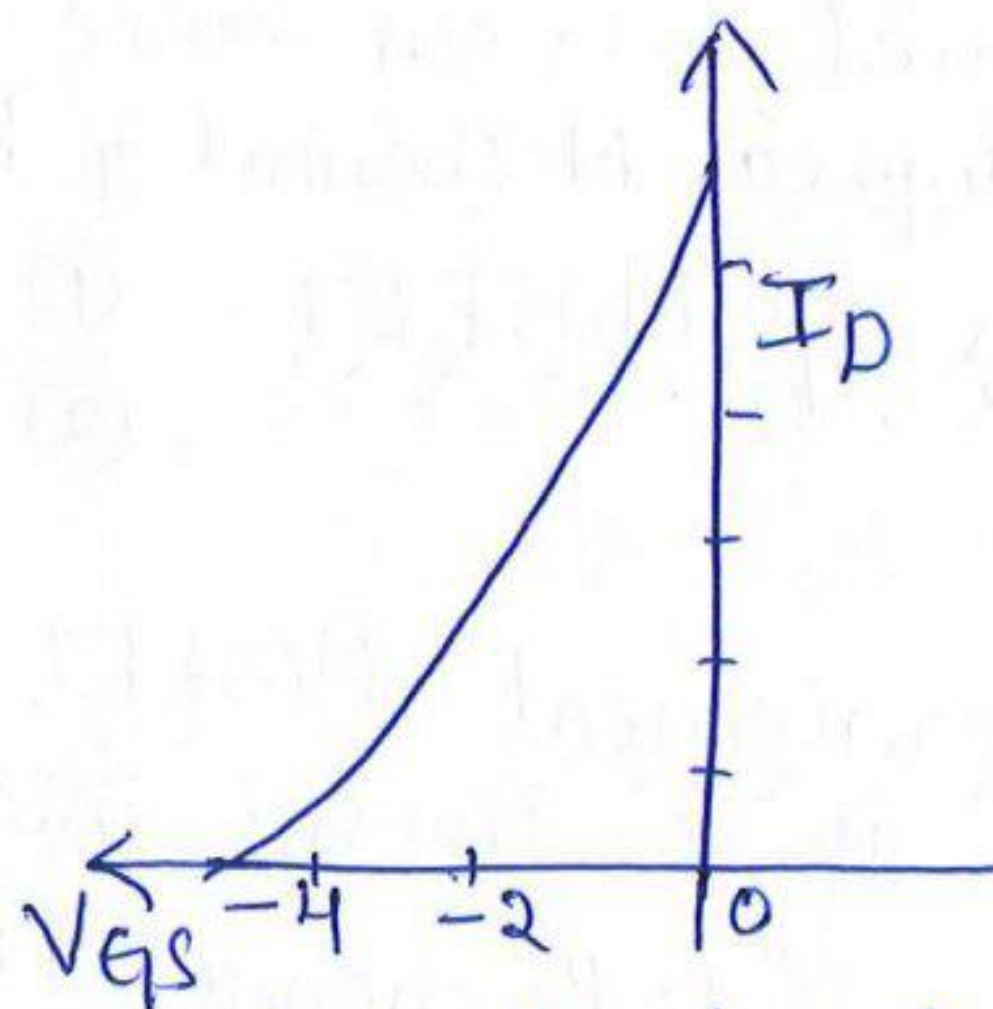


fig: Transfer characteristics

⇒ P-channel JFET:

- * In P-channel JFET, the channel is made of P-type material which is sandwiched between two N-type materials.
- * The battery connection is reversed i.e., drain is connected to the -ve terminal of the battery & source is connected to the +ve terminal.
- * Drain current flows from source to drain.
- * As the channel is made up of P-type material, the device is called P-channel JFET.

⇒ Applications of FET

- * Used for voltage amplification
- * Used in switching circuits
- * Used in operational amplifier as voltage variable resistors
- * Used in mixer circuits
- * Used in radio frequency amplifiers for FM devices
- * Used in computer memories because of its small size.

⇒ MOSFET (Metal Oxide Semiconductor FET)

- * It is a three terminal device also known as insulated gate field effect transistor (IGFET) as the gate of the MOSFET is insulated from the channel between the drain and the source.
- * Depending upon the type of channel (P or N) that is formed between source & drain, these are divided into 2 types as N-channel & P-channel

⇒ Types of MOSFET: (1) Enhancement MOSFET (2) Depletion MOSFET

⇒ Enhancement MOSFET:

It is divided into 2 types

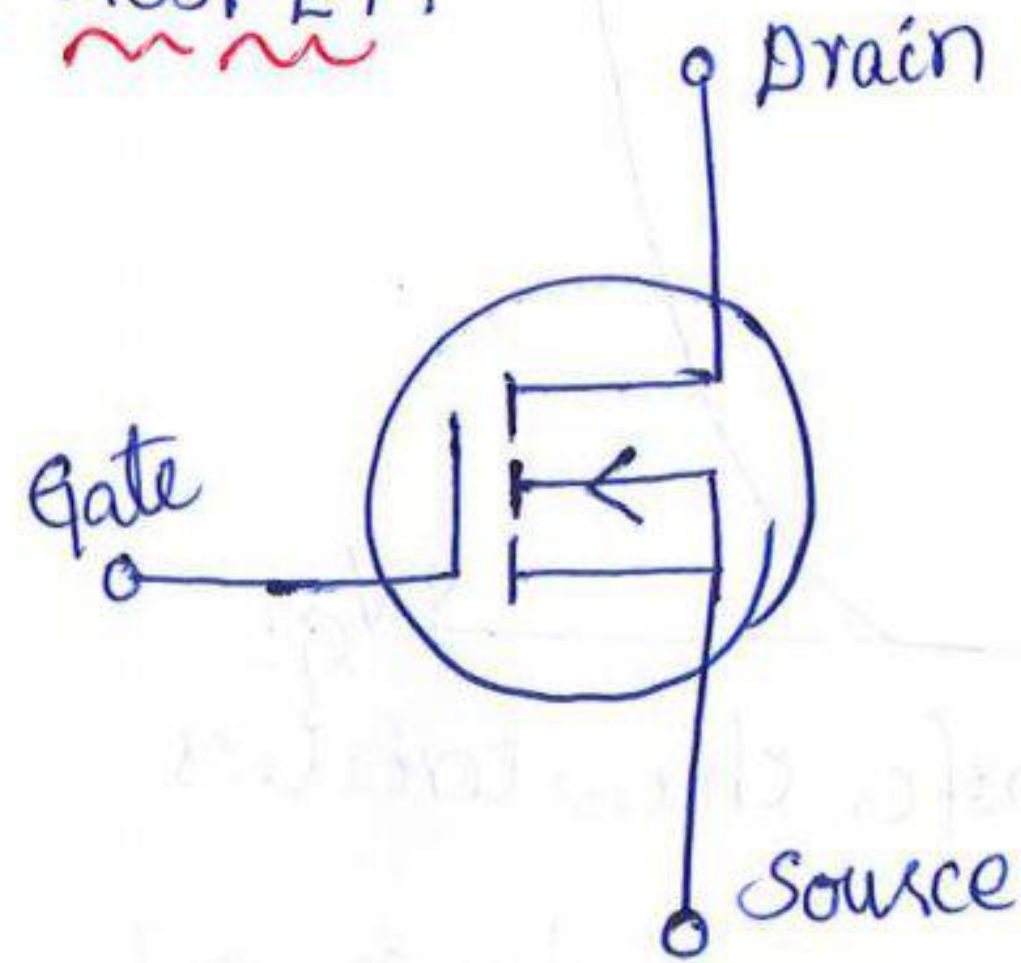
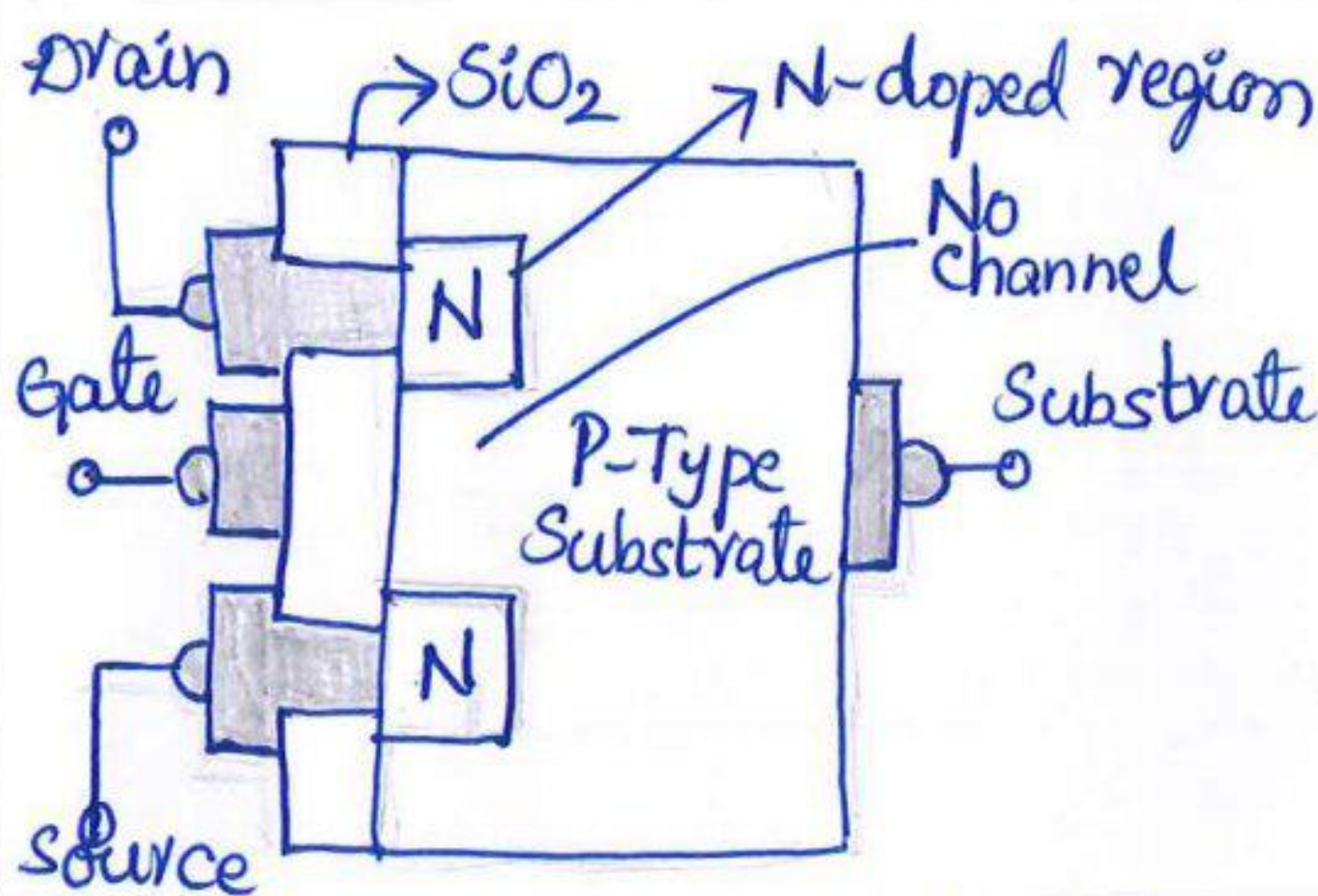
1. P-channel Enhancement MOSFET
2. N-channel Enhancement MOSFET

⇒ N-channel Enhancement MOSFET:

Construction:

- * Two heavily doped N-type materials are diffused into lightly doped P-type substrate.
- * The whole surface is coated with a layer of Silicon dioxide.
- * Metal contacts are made to take the terminals out.
- * The top terminal is called drain & the bottom terminal is called source.
- * Another metal or ohmic contact is made to take out the third terminal known as gate.
- * There is no link between source & drain (no physical channel)
- * The absence of channel is represented by a dashed line (---)

Structure & Symbol of N-channel MOSFET:



Operation of N-channel Enhancement MOSFET:

- * By shorting gate to source terminal, a sufficient drain to source voltage is applied by connecting the battery between drain & source terminals.
- * Source is connected to the -ve terminal & drain is connected to the +ve terminal.
- * No electrons move from source to drain due to absence of channel.
- * Even at a large +ve voltage at V_{DS} , current is zero.

- * A +ve Gate voltage is applied by connecting +ve terminal of V_{GS} to Gate and -ve terminal of V_{GS} to Source.
- * As the voltage V_{GS} increases, minority carriers in P-type i.e., electrons are attracted towards the wall side of SiO_2 layer.
- * At a large sufficient V_{GS} Voltage, large number of electrons are attracted towards the wall of Silicon dioxide layer due to which a link is established between Source & drain.
- * A channel is established or enhanced between Source & drain. Maximum number of carriers move from Source to drain which results in maximum drain current.

Characteristics of N-channel Enhancement MOSFET:

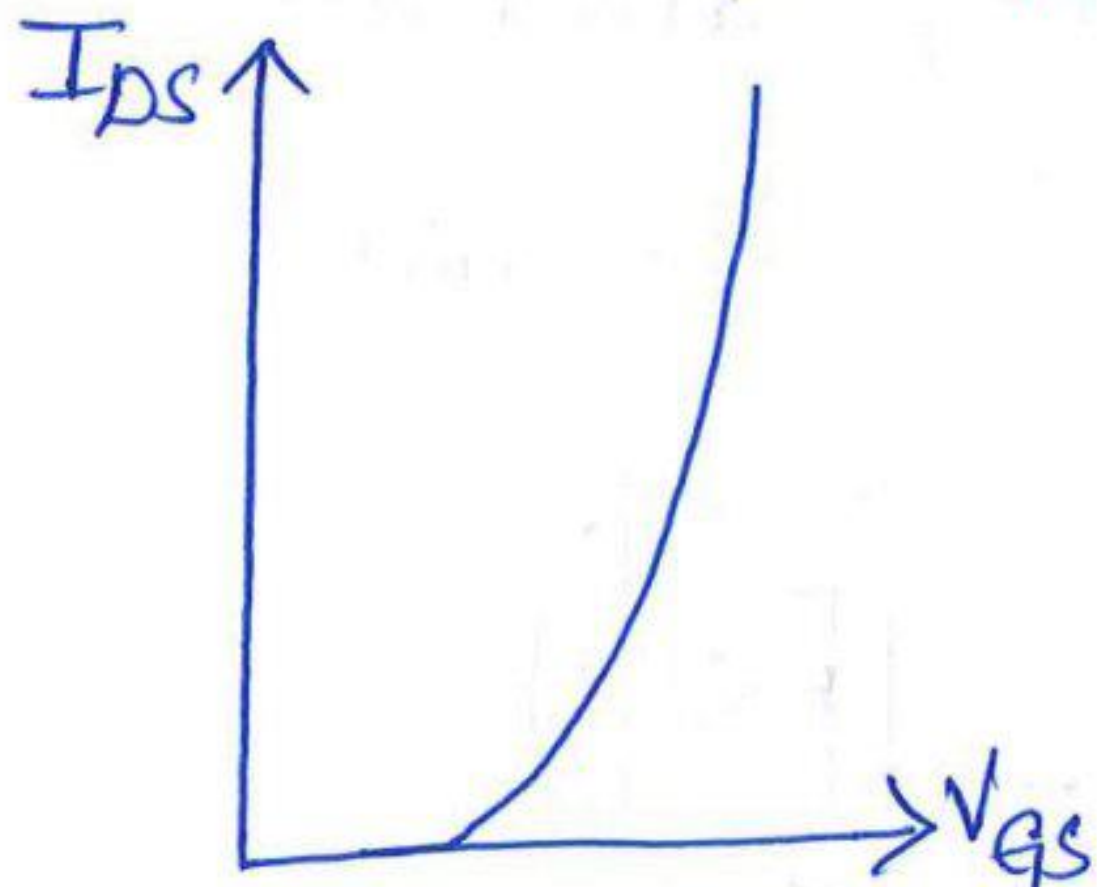


fig: Transfer characteristics

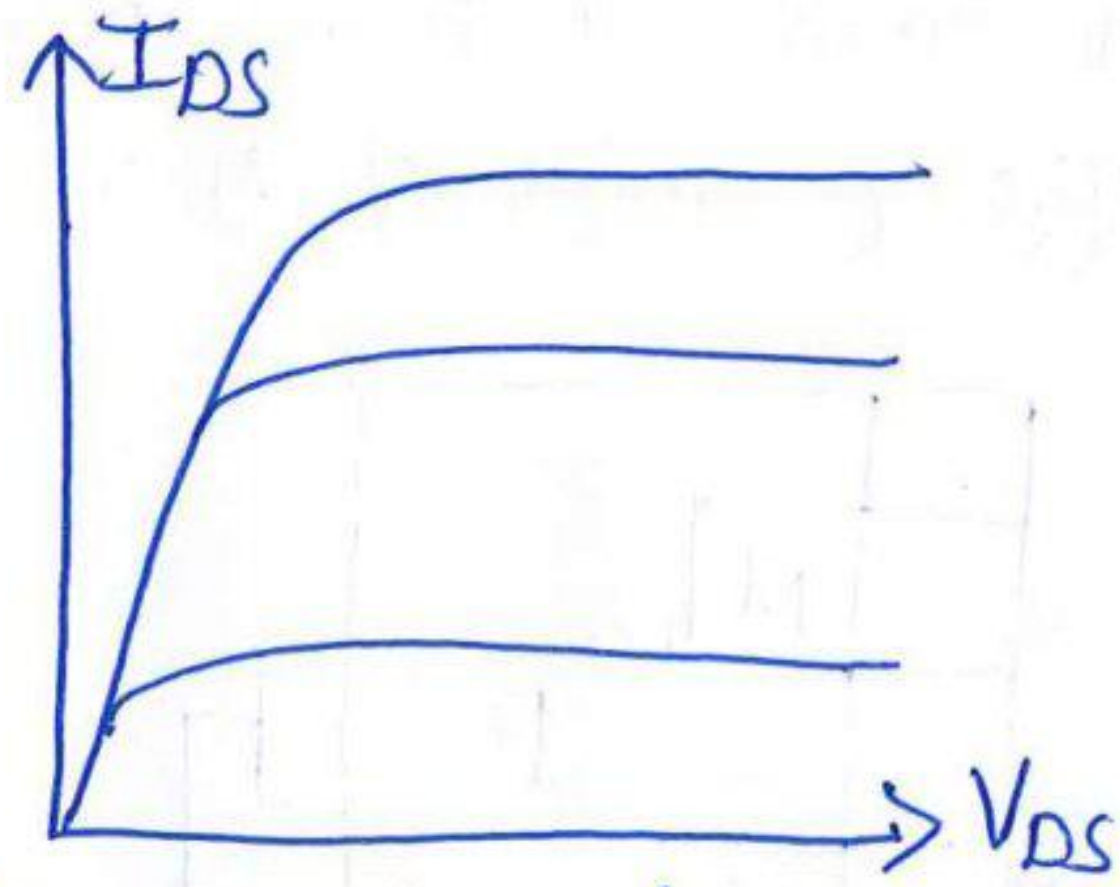


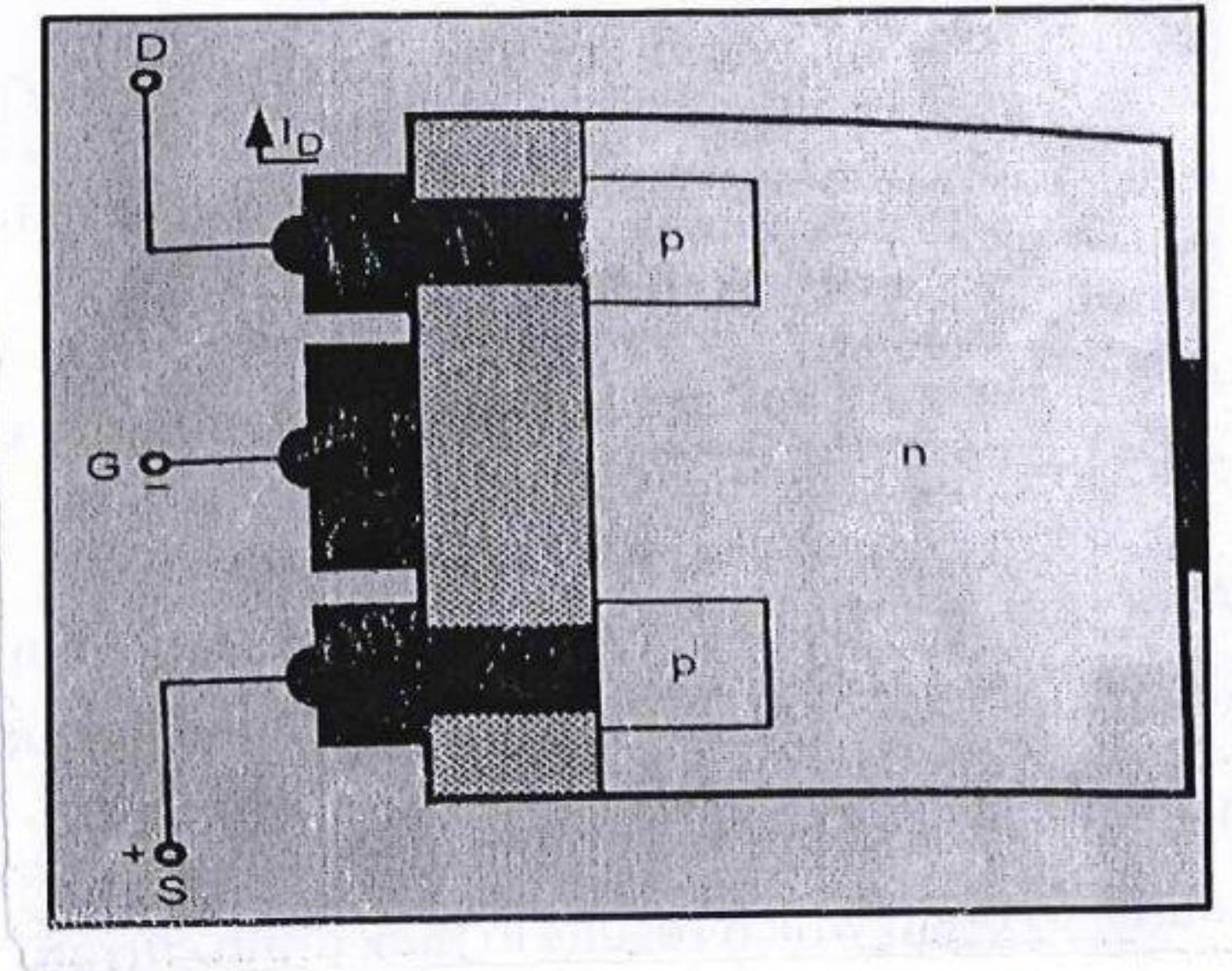
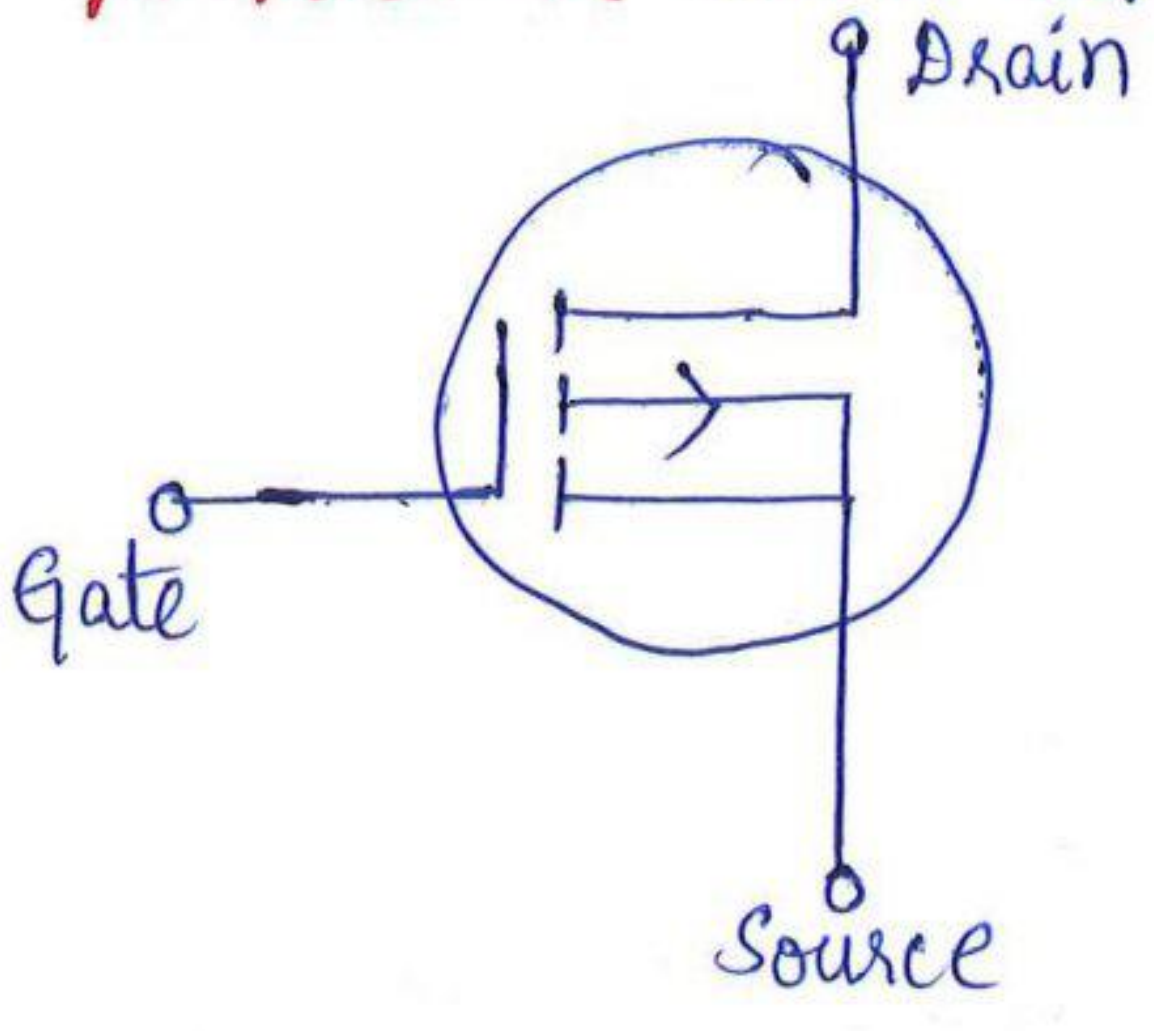
fig: O/p characteristics

⇒ P-channel Enhancement MOSFET:

Construction:

- * Two heavily doped P-type materials are diffused into lightly doped N-type substrate.
- * The whole surface is coated with a layer of Silicon dioxide.
- * Metal contacts are made to take out the terminals.
- * The top terminal is called drain & the bottom terminal is called source.
- * There is no link between source & drain (no physical channel).
- * Absence of channel is represented by a dashed line (---) in the symbol of P-channel Enhancement MOSFET.

Structure & Symbol of P-channel Enhancement MOSFET:



Operation:

- * The operation of P-channel Enhancement MOSFET is similar to that of N-channel Enhancement MOSFET with little difference.
- * Both V_{GS} and V_{DS} are -ve
- * Charge carriers are holes.

Characteristics of P-channel Enhancement MOSFET:

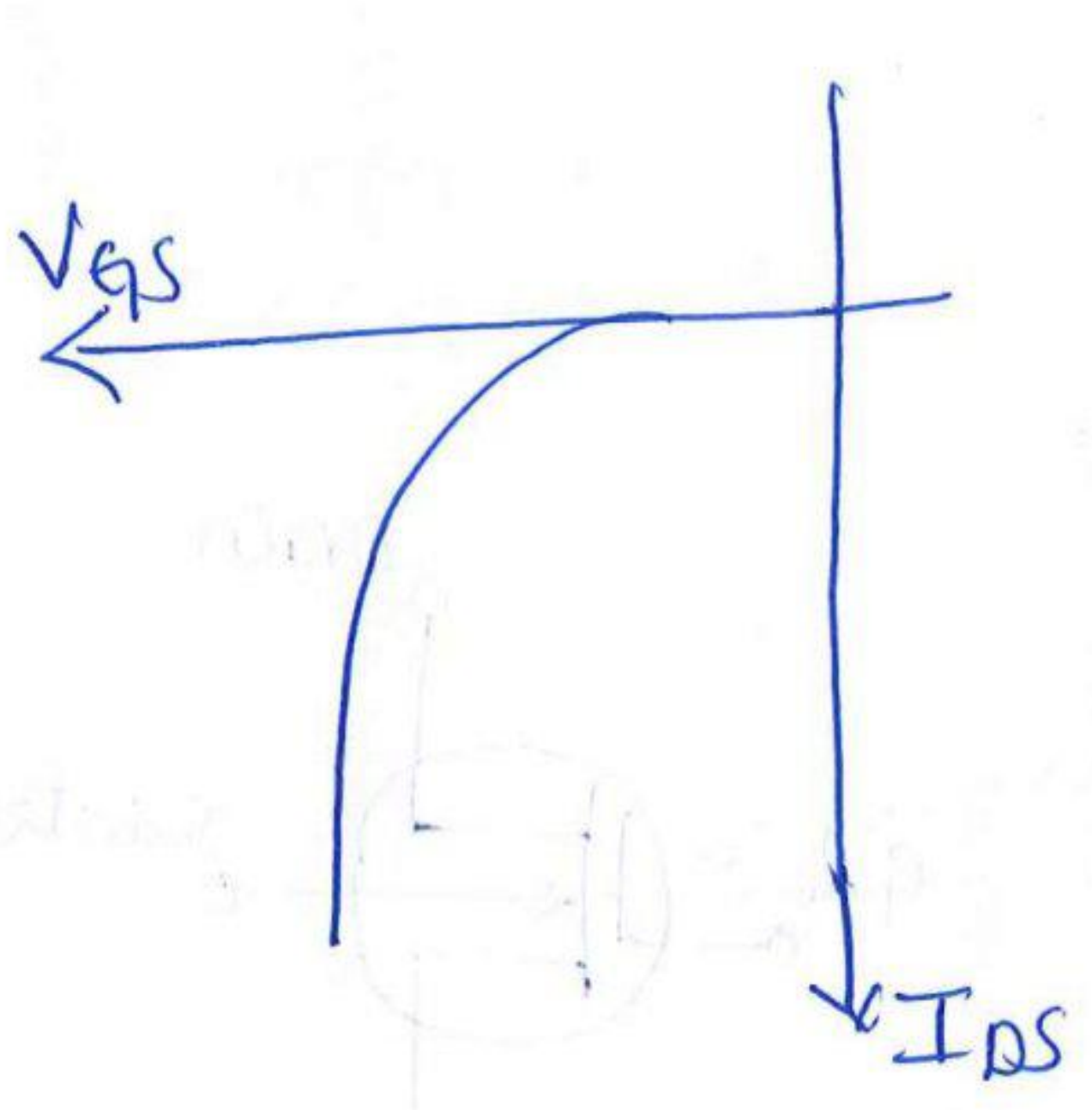


fig: Transfer characteristics

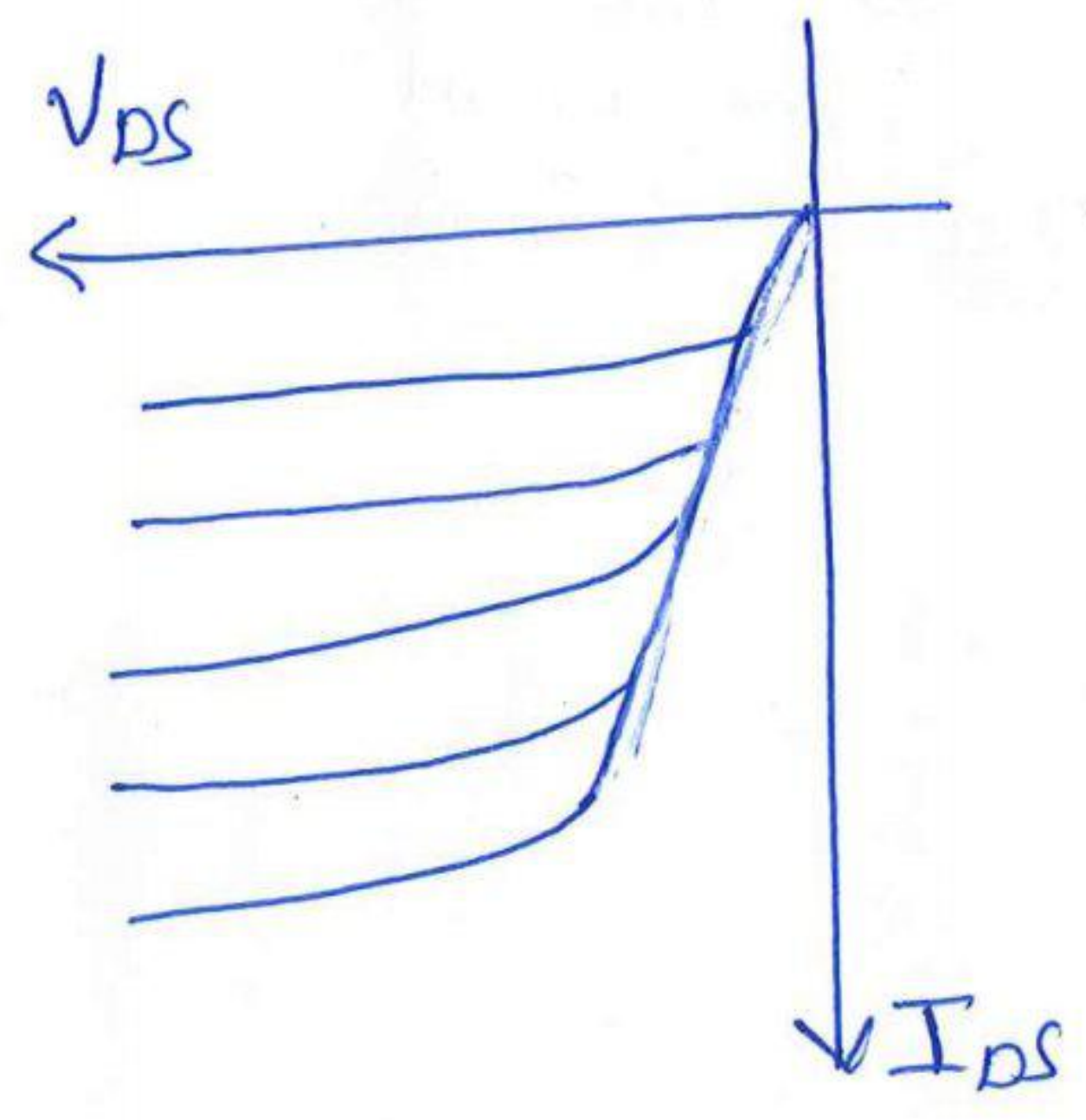


fig: o/p characteristics

⇒ Depletion Type MOSFET:

These are divided into 2 types.

① N-channel depletion type MOSFET

② P-channel depletion type MOSFET

⇒ N-Channel depletion type MOSFET:

Construction:

* Consider a lightly doped P-type substrate

* Add SiO_2 layer to the P-type substrate.

* Two highly doped N-type materials are diffused into lightly doped P-type substrate.

* Ohmic or metal contacts are made to take out the terminals.

* The top terminal is called drain & the bottom terminal is called source.

* The link between two N-type materials is called channel.

* As the channel is linking two N-type materials, it is called

N-channel MOSFET.

* In MOSFET, gate is separated from channel with SiO_2 layer.

Hence there is no direct electrical connection between gate & the channel.

Structure & Symbol

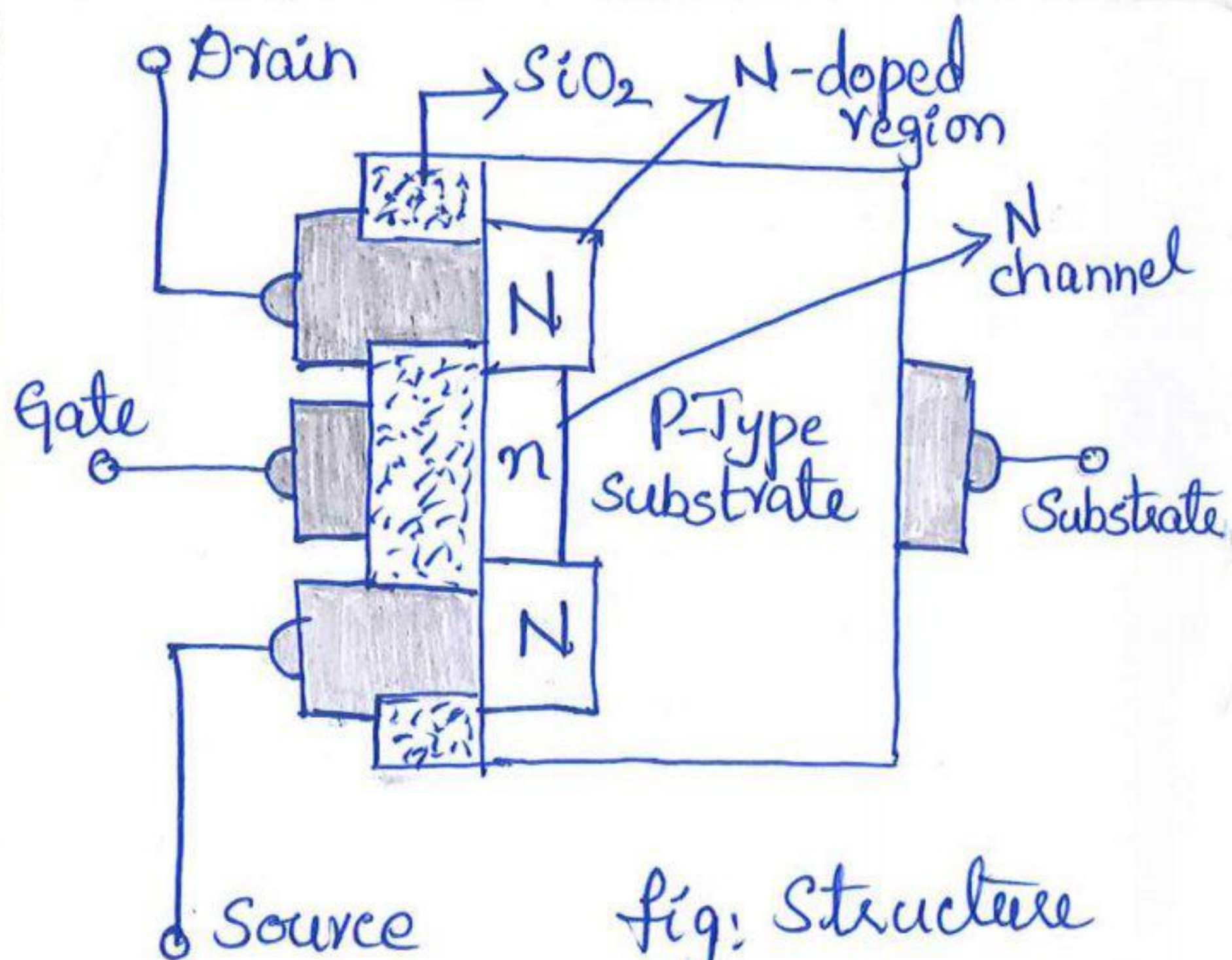


fig: Structure

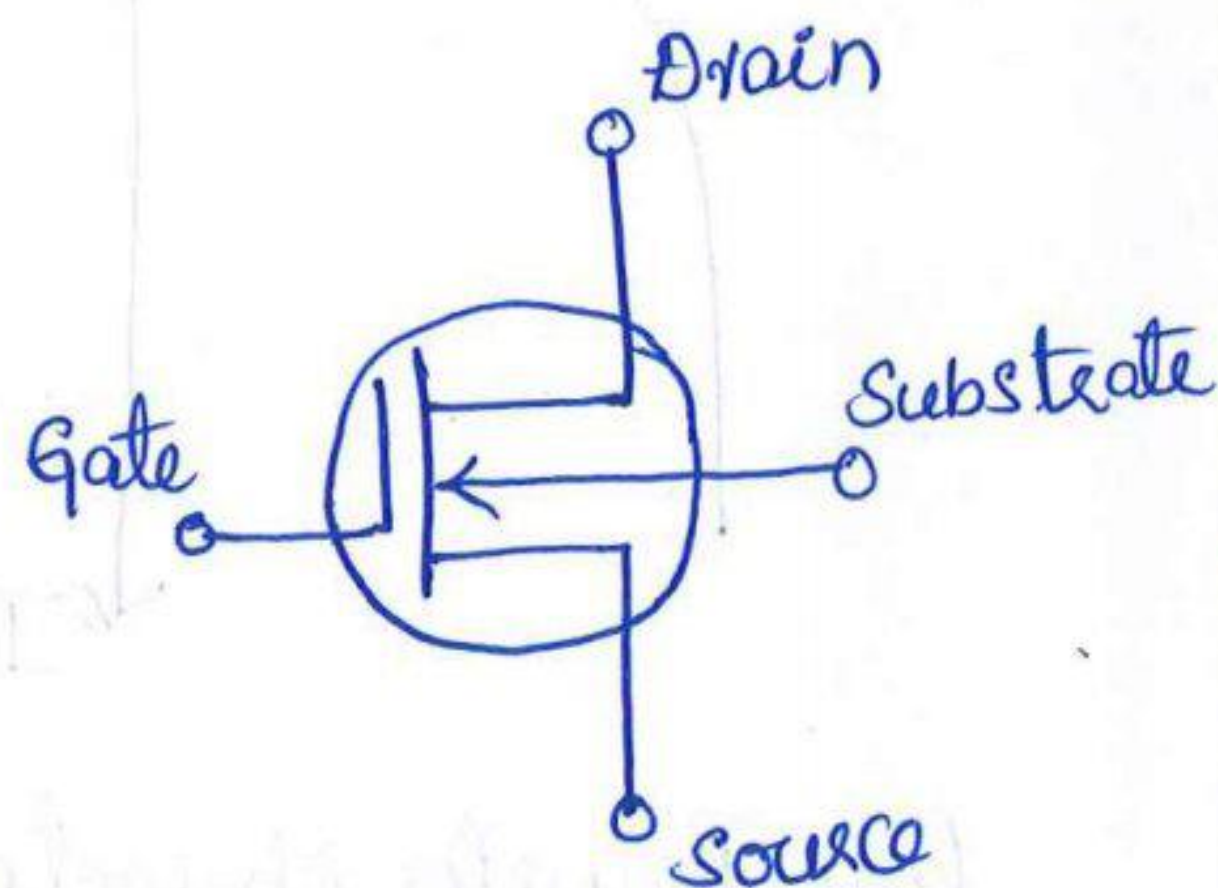
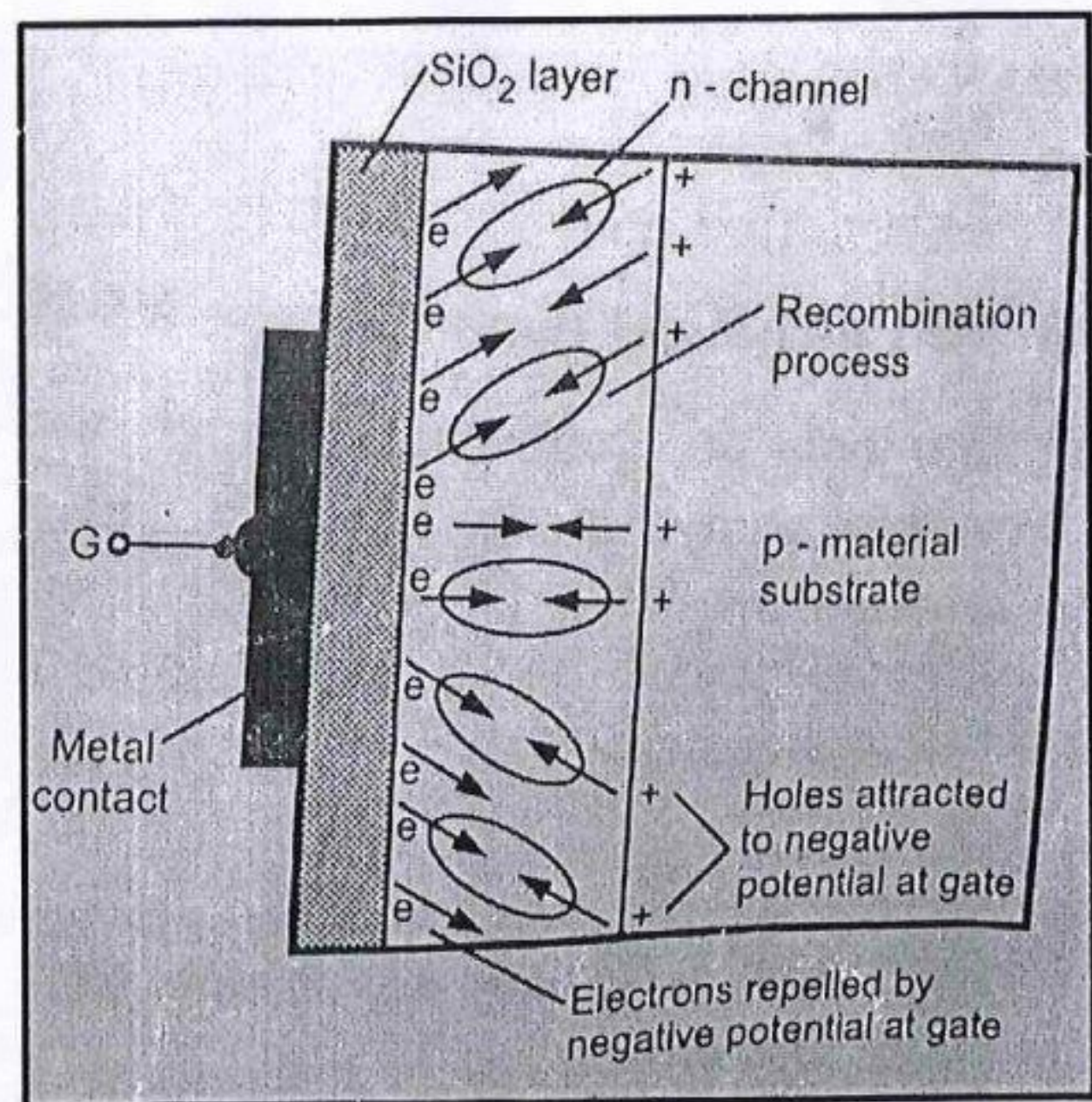
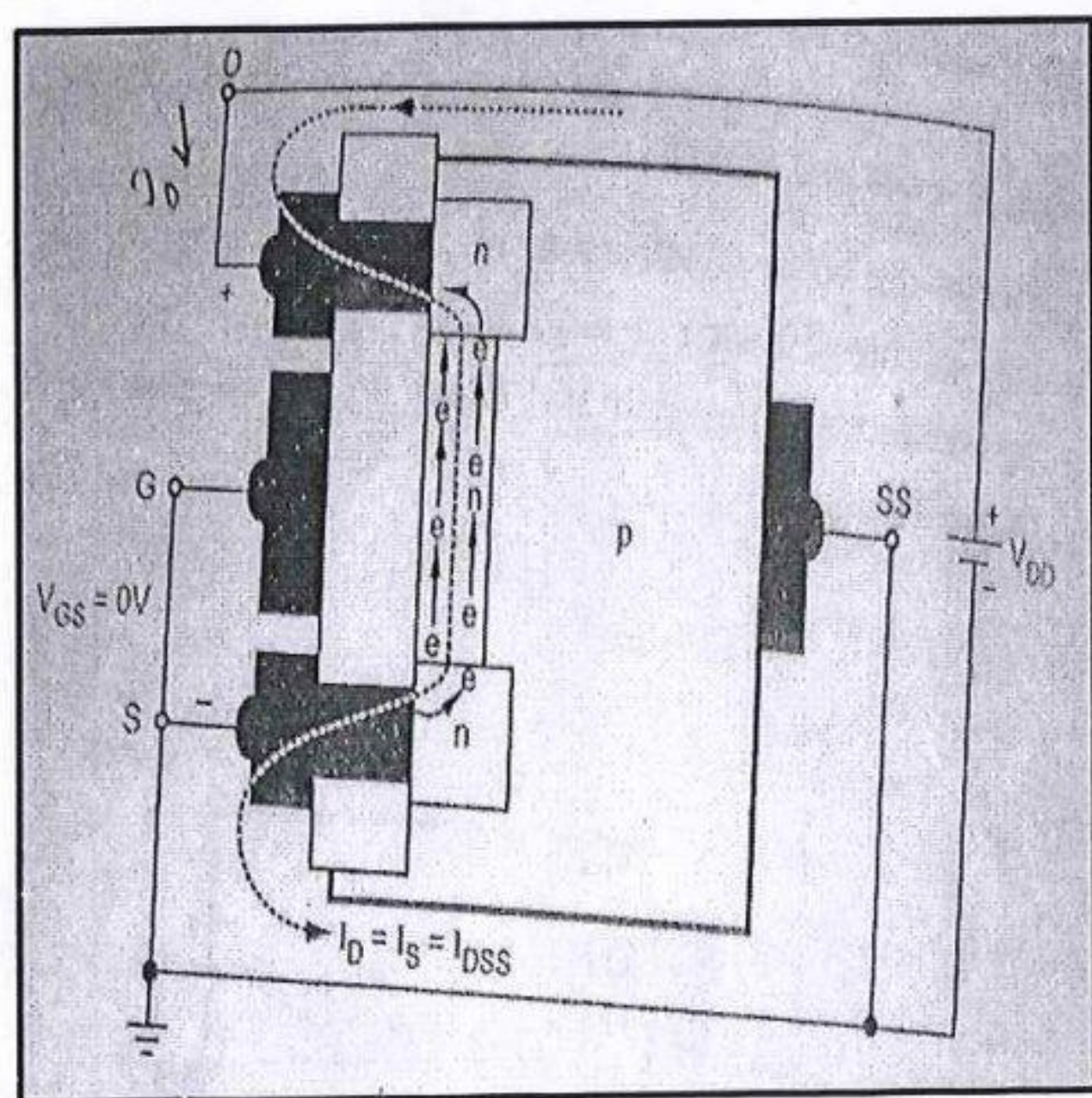


fig: Symbol

⇒ Operation of N-channel MOSFET:

- * Keeping V_{GS} as zero & applying voltage V_{DS} by connecting -ve terminal of V_{DS} to source & +ve terminal to drain, this makes the electrons to move from source to drain through the channel.
- * A further increase in voltage V_{DS} makes maximum electrons to move from source to drain through the channel causing maximum current
- * When voltage is applied between Gate & Source in such a way that -ve terminal is connected to Gate & +ve terminal of V_{GS} is connected to Source, holes from the p-type substrate get attracted towards the channel.
- * Repelled electrons recombine with attracted holes & create electron hole pairs.
- * Recombination reduces the number of free electrons in the n-channel for conduction reducing the drain current.
- * Greater the -ve voltage applied at the Gate, greater the depletion of n-channel electrons & level of drain current reduces.



⇒ P-channel depletion type MOSFET:

The entire procedure is similar to N-channel depletion type MOSFET with some changes like:

Replace P-type with N-type & N-type with P-type

Replace holes with electrons & electrons with holes.

Here there will be movement of holes

V_{GS} voltage is +ve

⇒ Applications of MOSFET:

1. Used for switching & amplifying electronic signals.
2. Used as an inverter.
3. Used as high frequency amplifier
4. Used as phase shift oscillator
5. Used in Switch Mode Power Supply.

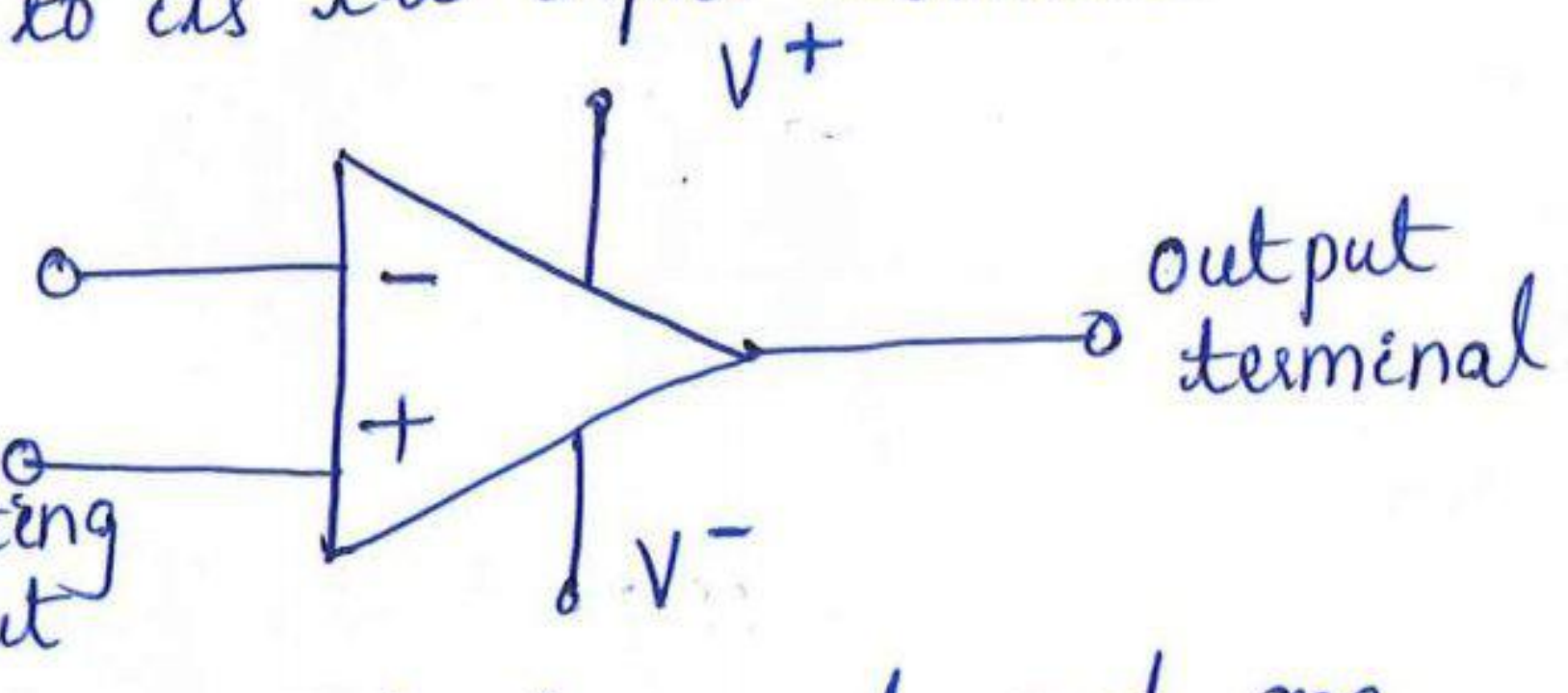
⇒ Operational Amplifier:

Definition: An operational amplifier is a high gain electronic voltage amplifier with a differential input and single ended o/p.

* It is a differential amplifier that amplifies the difference of voltages applied to its two input terminals.

Circuit Symbol:
Inverting input

Non-inverting input



* An op-amp has two input terminals and one output terminal

* It also has 2 voltage supply terminals.

* The term Op-Amp is used to denote an amplifier that performs various operations like addition, subtraction, differentiation, integration etc.

* The terminal with a (-) sign is called inverting input terminal.

* When an input signal is applied at the inverting terminal, the o/p is 180° out of phase with the i/p signal.

* The terminal with a (+) sign is called non-inverting input terminal.

* When an input signal is applied at the non inverting terminal, the o/p is in phase with the i/p signal.

⇒ Characteristics of an ideal operational Amplifier:

1. Infinite open loop Voltage gain

2. Infinite input impedance

3. Zero Output impedance

4. Bandwidth is infinity

5. An ideal op-amp draws no current at both the input terminals

6. Zero offset

⇒ Inverting Amplifier:

* Inverting amplifier is an op-amp in which the output is given as feedback to the inverted terminal of the input using a feedback resistor.

* In an inverting amplifier,

if the input signal is +ve, then the o/p is -ve,

if the input signal is -ve, the o/p is +ve.

* Input signal V_i is applied to the inverting terminal through R_1

Non-inverting terminal of the op-amp is grounded.

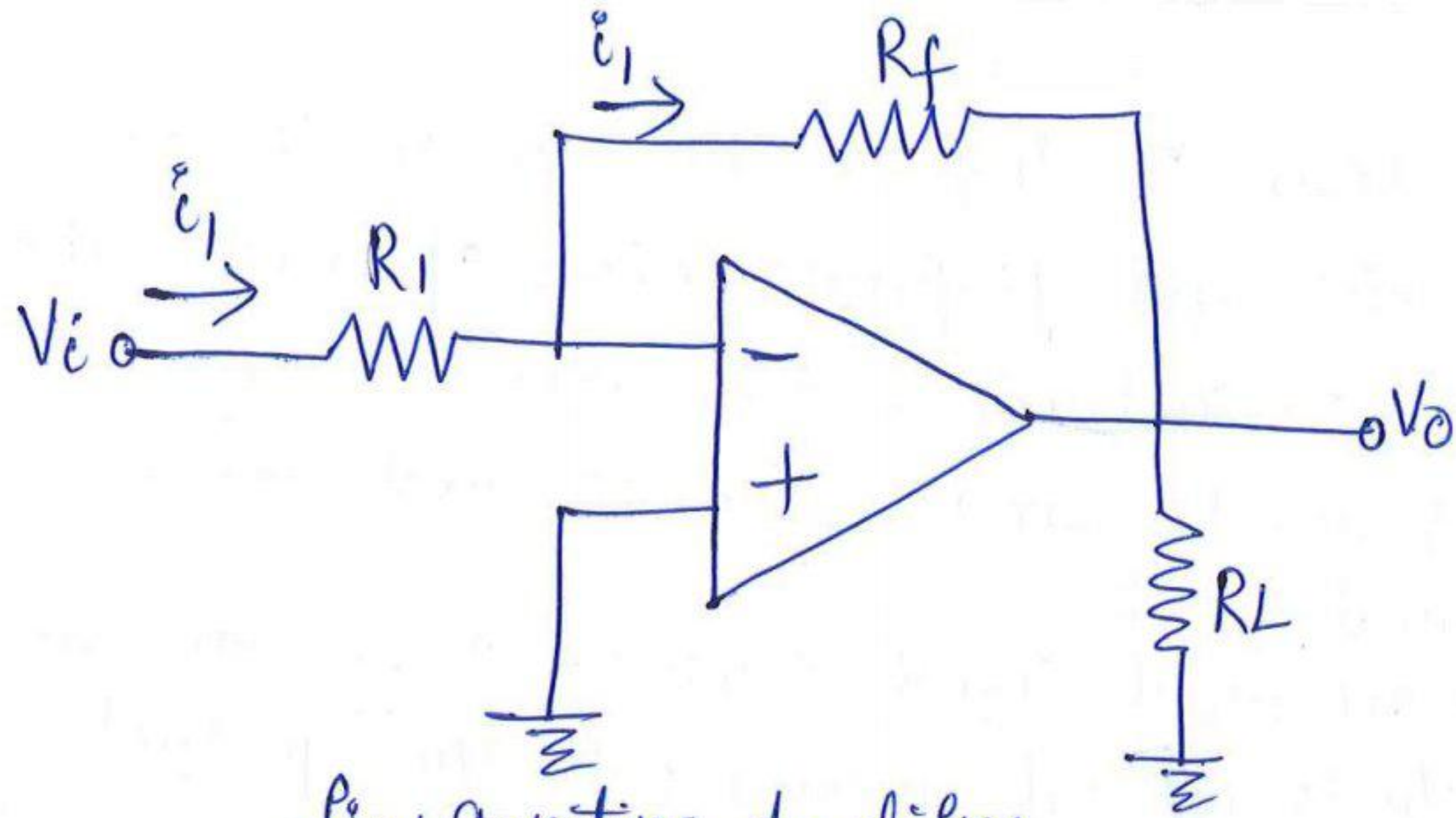


fig: Inverting Amplifier

* Current i_1 flowing through R_1 is

$$i_1 = \frac{V_i}{R_1} \quad \text{--- (1)}$$

ohms law

$$V = IR$$

$$I = V/R$$

* As op-amp draws no current, all the current flowing through R_1 must flow through R_f

* Output Voltage, $V_o = -i_1 R_f$ --- (2)

Substitute equ (1) in equ (2)

$$V_o = -\left(\frac{V_i}{R_1}\right) R_f = -V_i \frac{R_f}{R_1}$$

$$\boxed{V_o = -V_i \frac{R_f}{R_1}} \quad ; \quad \frac{V_o}{V_i} = -\frac{R_f}{R_1}$$

* Gain of the inverting amplifier is

$$\boxed{A = \frac{V_o}{V_i} = -\frac{R_f}{R_1}}$$

* The -ve sign indicates a phase shift of 180° between V_i & V_o .

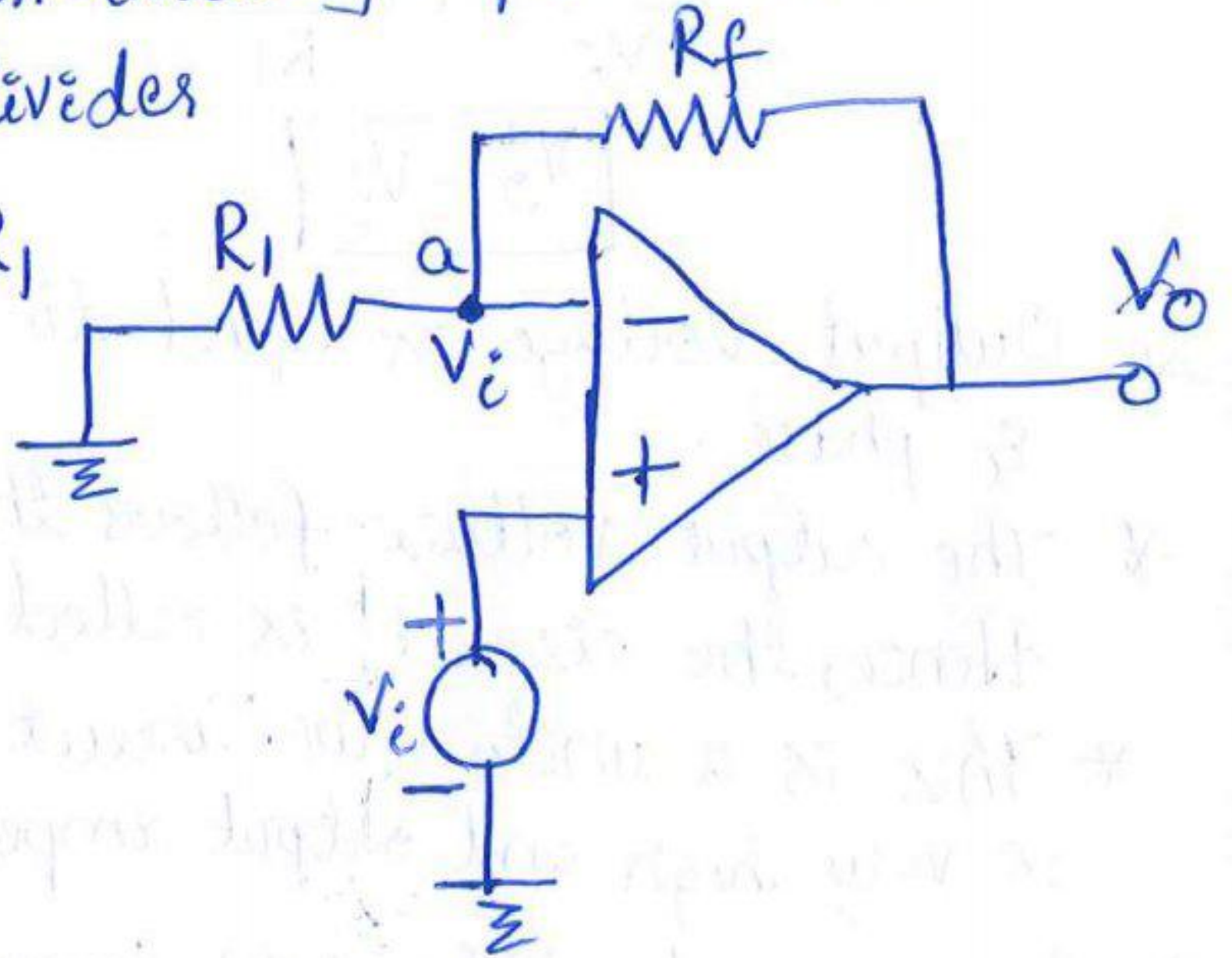
⇒ Non-Inverting Amplifier:

A non-inverting amplifier amplifies the input signal and the output signal is in-phase with the input signal.

- * A non inverting amplifier behaves like a voltage follower circuit.
- * If a signal is applied at the non-inverting input terminal the circuit amplifies without inverting the input
- * It is a negative feedback system as output is fed back to the inverting input terminal.
- * As the differential voltage V_d at the input terminal of the op-amp is zero, voltage at node 'a' is same as the i/p voltage applied at the non-inverting input terminal.
- * R_f & R_1 forms a potential divider

$$V_i = \frac{V_o}{R_1 + R_f} \cdot R_1$$

as no current flows into op-amp



$$\frac{V_o}{V_i} = \frac{R_1 + R_f}{R_1} = \frac{R_1}{R_1} + \frac{R_f}{R_1}$$

$$\frac{V_o}{V_i} = 1 + \frac{R_f}{R_1}$$

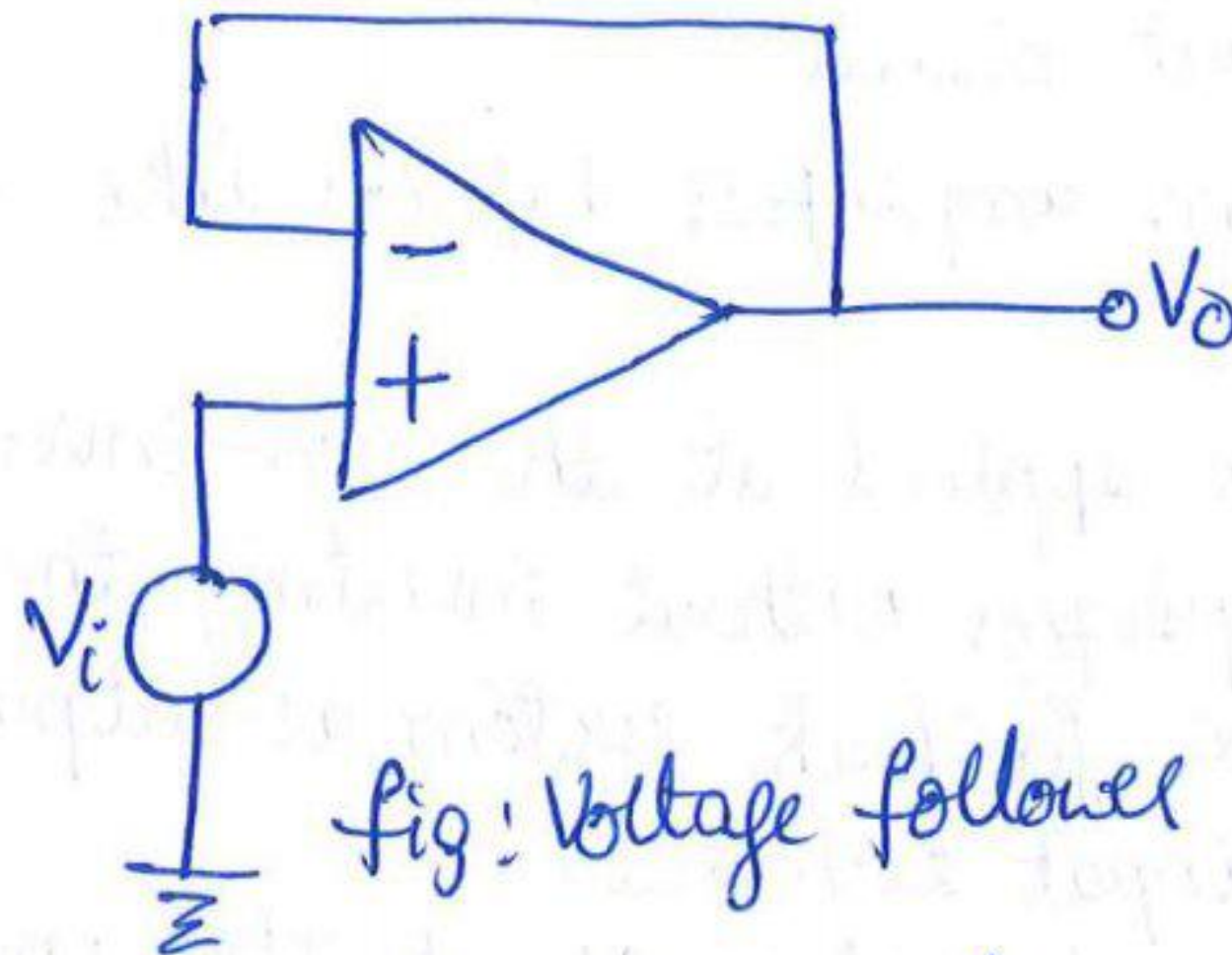
* Voltage Gain for the Non-inverting Amplifier is

$$A_{CL} = \frac{V_o}{V_i} = 1 + \frac{R_f}{R_1}$$

$$V_o = \left(1 + \frac{R_f}{R_1}\right) V_i$$

⇒ Voltage Follower:

In the non-inverting amplifier, if $R_f = 0$ & $R_1 = \infty$ we get voltage follower circuit



Substitute $R_f = 0$ & $R_1 = \infty$ in the equation

$$\frac{V_o}{V_i} = 1 + \frac{R_f}{R_1} = 1 + \frac{0}{\infty} = 1$$

$$\boxed{V_o = V_i}$$

- * Output voltage is equal to input voltage both in magnitude & phase.
- * The output voltage follows the input voltage exactly. Hence, the circuit is called a voltage follower.
- * This is a unity gain circuit whose input impedance is very high and output impedance is zero.

⇒ Summing Amplifier (or) Summer:

- * Op-amp may be used to design a circuit whose output is the sum of several input signals. Such a circuit is called summing amplifier.
- * Consider an inverting summing amplifier with three input voltages V_1, V_2 and V_3 , three input resistors R_1, R_2, R_3 and a feedback resistor R_f .

$$V_o = -V \left(\frac{R_f}{R_1} + \frac{R_f}{R_2} + \frac{R_f}{R_3} \right)$$

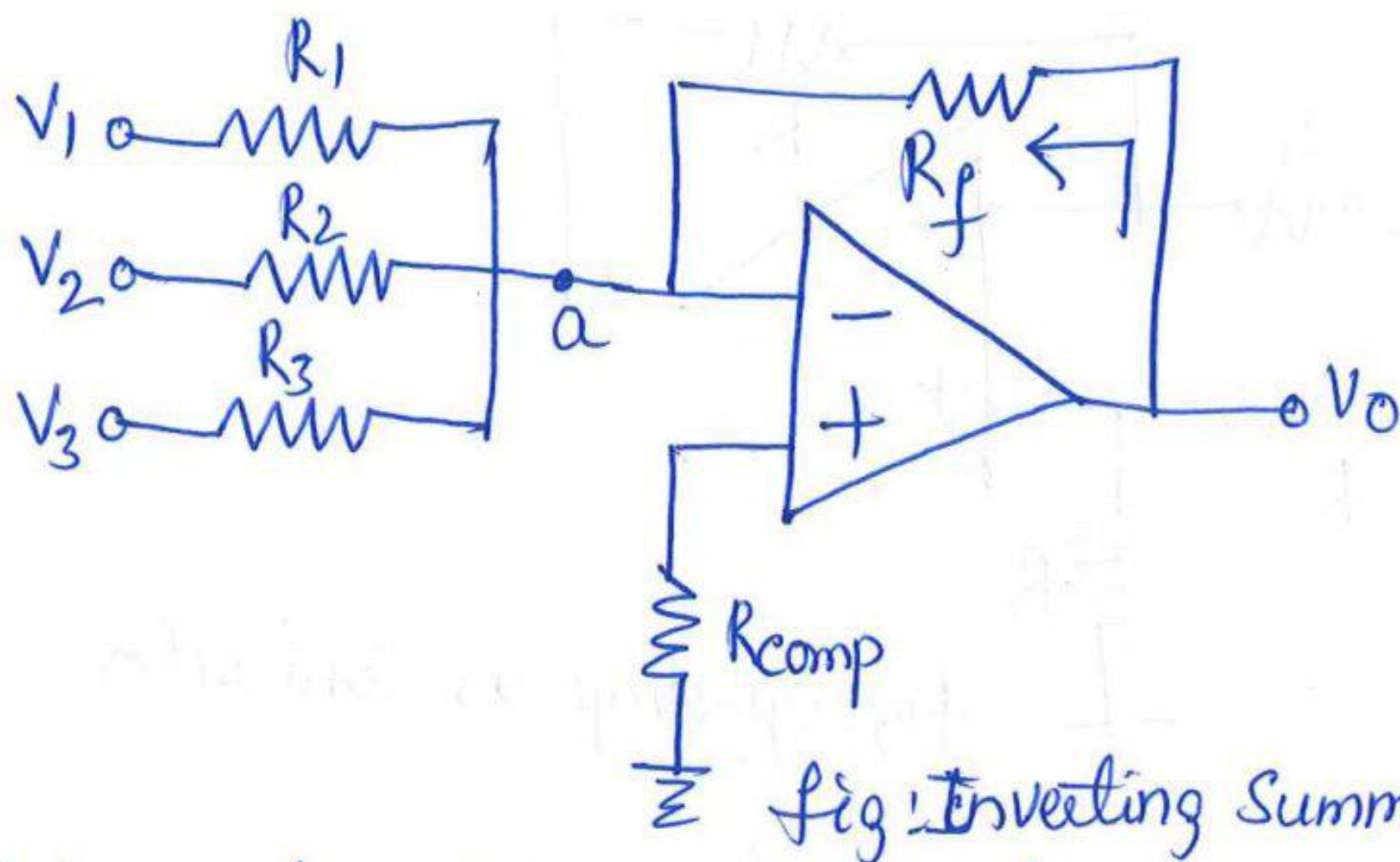


Fig: Inverting Summing Amplifier

* Nodal equation by KCL at node 'a' is

$$\frac{V_1}{R_1} + \frac{V_2}{R_2} + \frac{V_3}{R_3} + \frac{V_0}{R_f} = 0$$

$$\frac{V_0}{R_f} = -\left(\frac{V_1}{R_1} + \frac{V_2}{R_2} + \frac{V_3}{R_3}\right)$$

$$V_0 = -\left[\left(\frac{R_f}{R_1}\right)V_1 + \left(\frac{R_f}{R_2}\right)V_2 + \left(\frac{R_f}{R_3}\right)V_3\right]$$

* In special case, when $R_1 = R_2 = R_3 = R_f$

$$V_0 = -(V_1 + V_2 + V_3)$$

Output V_0 is the inverted sum of input signals.

⇒ Subtractor:

* A basic differential amplifier can be used as a subtractor.

* If all the resistors are equal in value, then the o/p can be derived using superposition principle.

* To find the output V_{o1} due to V_1 alone, make $V_2 = 0$.

then the circuit becomes a non-inverting amplifier

having input voltage $V_1/2$ at the non-inverting terminal

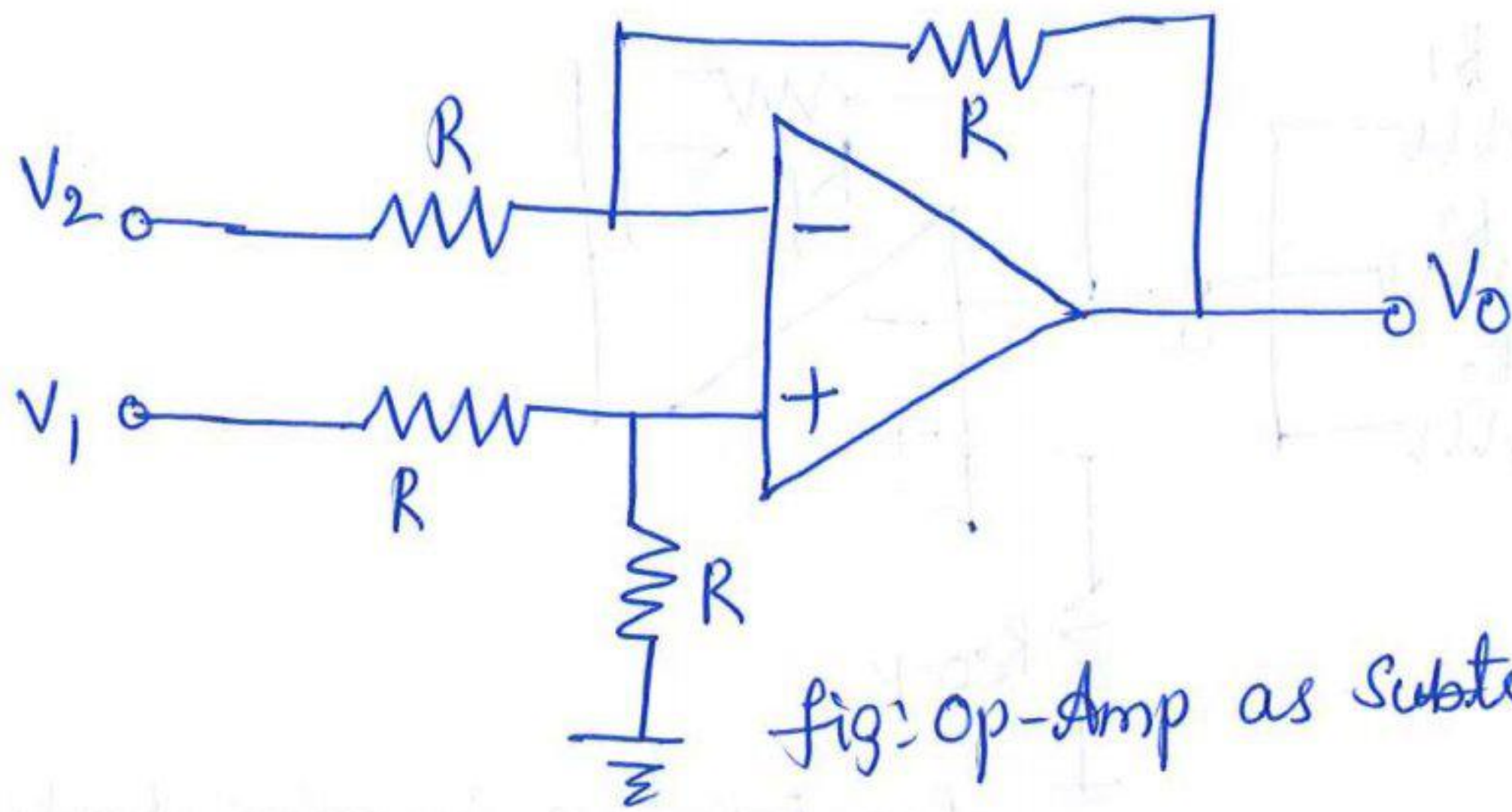


fig: op-amp as subtractor

* Making $V_2=0$, o/p V_{o1} due to V_1 alone is

$$V_{o1} = \frac{V_1}{2} \left(1 + \frac{R}{R}\right) = \frac{V_1}{2} (1+1) = \frac{V_1}{2} \times 2$$

$$V_{o1} = V_1$$

* Making $V_1=0$, o/p V_{o2} due to V_2 alone is

$$V_{o2} = -V_2$$

* Output voltage V_o due to both the inputs is

$$V_o = V_{o1} + V_{o2} = V_1 - V_2$$

$$V_o = V_1 - V_2$$

output is the difference of the input signals.

⇒ Block diagram of an operational amplifier:

* An op-amp has 4 cascaded blocks.

* The first 2 stages are cascaded differential amplifiers and these provide high gain and high input resistance.

* The third stage acts as a buffer and level shifter.

* The output stage is designed to provide low o/p impedance.

* The output stage is designed to provide low o/p impedance.



fig: Block diagram of an op-amp

(a) Differential Amplifier:

A differential amplifier amplifies the difference between two input signals.

It will suppress undesired noise & provides high gain.

It also provides high input resistance.

(b) Buffer and Level Shifter

Buffer is an emitter follower whose input impedance is very high.

Level shifter adjusts the dc voltages so that output voltage is zero for zero inputs.

(c) Output Stage:

The output stage is designed to provide low output impedance.

to differentiate amplifier

A differentiated amplifier amplifies the difference

between two input signals

It will produce maximum output if the two input signals

are 180 degrees out of phase

As buffer and gain buffer

buffer is an emitter follower whose input impedance

is very high

and output is like a voltage divider

output is low for two inputs

output stage

The output stage is designed to provide low output

impedance

Communication Systems:

* The function of communication system is to transfer information from one place to another through communication channel.

* Telephone was invented in 1876. Speech was converted into electrical signals which were sent using wires to long distances.

* The electrical signals received through the receiver set was converted back to sound signal.

⇒ Elements of Communication System:

* The basic components of a communication system are:

- 1. Transmitter
- 2. Communication channel
- 3. Receiver

* Transmitter: Message or information signal is given as an input to the transmitter.

Transmitter is an electronic device that converts the message into signals which can be easily transmitted through the communication channel.

Ex of transmitter is microphone which converts speech into electrical signal.

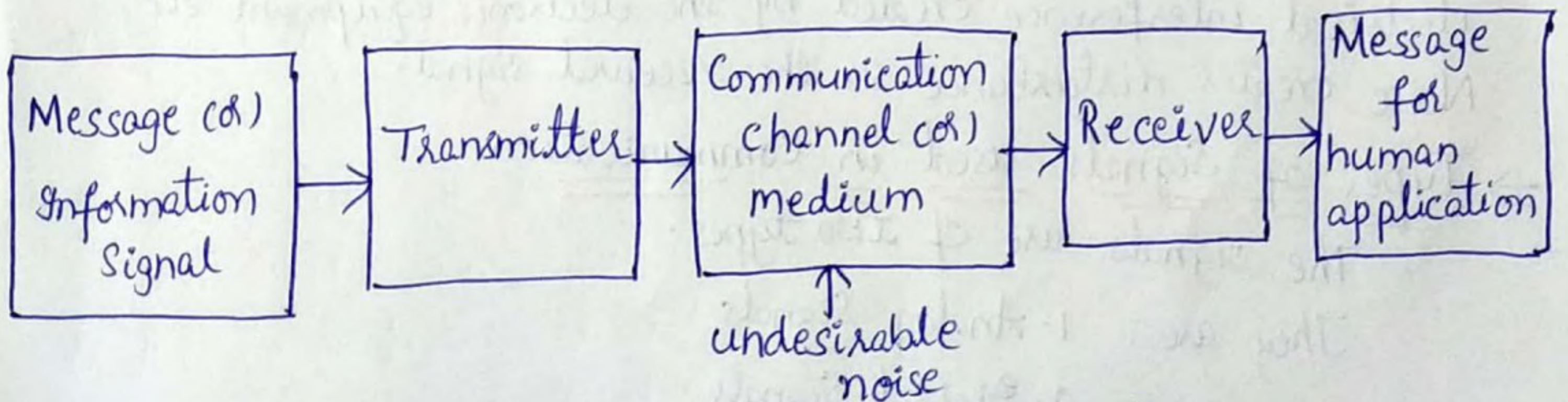


Fig: Block diagram of a communication system

Communication channel:

The communication channel is used to send the message from one place to another.

- * Communication channel can be a pair of wires, fibre-optic cable or it can be wireless.
- * Wireless Communication is referred as radio.
- * Signals are communicated from one place to another for long distances using electromagnetic field waves.
- * Signals are converted into electric and magnetic fields that propagate easily over long distances in a second.
- * Radio Communication is wireless & it does not require electrical connection between the transmitter and the receiver.

Receives:

- * When the signal is transmitted, it gets reduced in magnitude. Hence, it is necessary to amplify the message signal both at the transmitter & receiver.
 - * At the transmitting end, one antenna transmits the signal through using a high frequency carrier wave.
 - * The receiver picks up the signal through another antenna & then separates the signal from the carrier wave.
 - * Receiver converts the signal back to its original form. [Separates the message signal from the carrier]
- Noise is an undesirable electric energy which enters the communication channel due to many reasons. Noise comes from the atmosphere in the form of radiation energy, electrical interference created by the electronic equipment etc., Noise creates disturbance in the received signal.

→ Types of Signals used in Communication:

The signals are of two types.

- They are:
1. Analog Signals
 2. Digital Signals

1. Analog Signals: Analog signals are continuous signals.
 Analog signal is a continuously varying voltage or current.
 Ex: When we speak using a microphone, an analog electrical voltage signal which continuously varies with time is generated.

2. Digital Signals: Digital signals are discrete signals.
 Data used in computers is digital where binary codes are used to represent numbers, letters, symbols etc.,

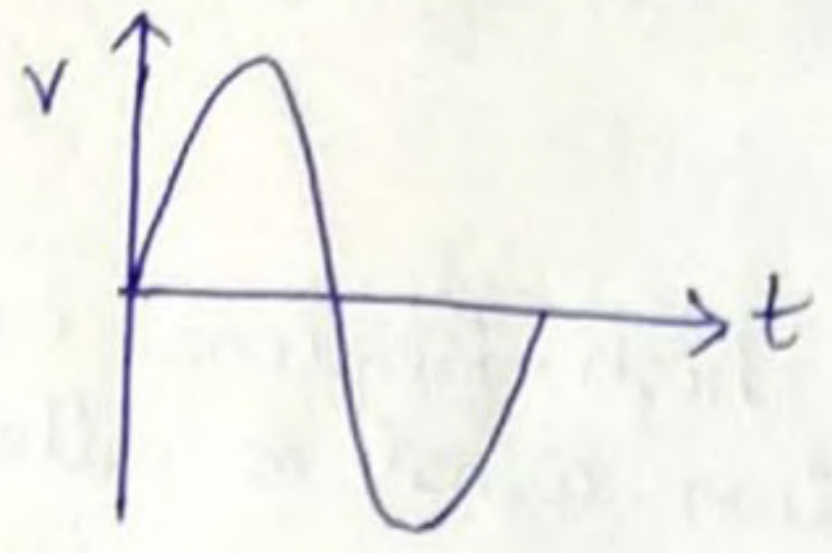


fig: Analog Signal

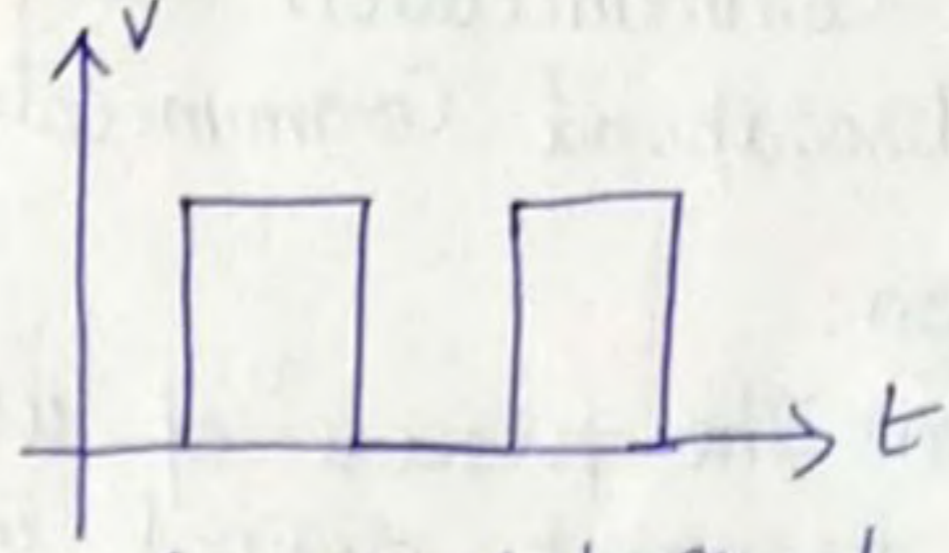


fig: Digital Signal

⇒ Direction of Communication:

Depending on whether the communication is one way or both ways, it is classified as

- ① Simplex
- ② Half duplex
- ③ Full duplex

Simplex Systems:

The systems which simply transmit & do not receive any information are called simplex systems.
 Ex: Radio & TV communicate in one direction only.

Half Duplex Systems:

When communication is done in both the directions it is called duplex communication system.
 Ex: Telephone communication is a two-way communication.

* Communication systems which are bidirectional but not simultaneous, i.e., communication is possible in one direction at a time. Such systems are called half duplex systems.

Full Duplex Systems:

In full duplex systems, the communication between sender & receiver can occur simultaneously.

⇒ Baseband Signals:

The original information signals, either analog or digital are called baseband signals.

⇒ Baseband Transmission System:

When baseband signals are directly communicated through the channel, it is called baseband transmission system.

⇒ Broadband Communication:

Communication using a carrier frequency wave is called broadband communication.

⇒ Modulation:

* Definition: The process of using a high-frequency carrier wave to transmit original information signal is called modulation.

* Modulation is the process of placing the information signal which is of low frequency over a high frequency carrier signal for long-distance transmission.

⇒ Types of Modulation:

Different types of modulation are:

1. Amplitude Modulation
2. Frequency Modulation
3. Phase Modulation

* The message or information signal which is of low frequency is made to ride on the carrier signal having high frequency through a process called modulation.

Ex: The frequency of information signal (voice) ranges from 300 to 3000 Hz. They are used to modulate a carrier of higher frequency.

The modulated electromagnetic wave travels through the free space at a high speed. The receiver picks up this modulated electromagnetic wave. This is then converted into electrical signal. The carrier & the information signals are separated through a process called demodulation & the original message is recovered at the user end.

⇒ Need for Modulation:

The advantages of using modulation technique are:

- ① Reduction in the height of the antenna.
- ② Increase in the range (distance) of communication.
- ③ Avoidance of mixing up of information signals.
- ④ Improvement of quality of reception of the information sent.
- ⑤ Multiplexing of signals is possible.
- ⑥ Allowing adjustment of bandwidths.

⇒ Problem: Calculate the height of the antenna required for baseband transmission of a signal of frequency 3kHz. Also calculate the height of the antenna required when a carrier frequency of 2MHz is used for transmitting a modulated signal.

Solution: Minimum height of the antenna required = One fourth of the wavelength. i.e., $h = \frac{\lambda}{4}$ where

Wavelength $\lambda = c/f$; $c = \text{velocity of light} = 3 \times 10^8$; $f = 3 \text{ kHz} = 3 \times 1000 \text{ Hz}$

$$h = \frac{\lambda}{4} = (c/f) \cdot \frac{1}{4}$$

$$h = \frac{3 \times 10^8}{3 \times 1000} \times \frac{1}{4} = 25 \text{ km}$$

$$\text{When } f = 2 \text{ MHz} = 2 \times 10^6; h = \frac{\lambda}{4} = \frac{c}{f} \times \frac{1}{4} = \frac{3 \times 10^8}{2 \times 10^6} \times \frac{1}{4} = 37.5 \text{ m}$$

⇒ Electromagnetic Spectrum:

<u>Frequency Range</u>	<u>Name of the frequency</u>	<u>Wavelength</u>
3 - 300 Hz	Extremely low frequencies (ELF)	$10^7 - 10^6 \text{ m}$
300 - 3000 Hz	Voice Frequencies (VF)	$10^6 - 10^5 \text{ m}$
3 - 30 kHz	Very low frequencies (VLF)	$10^5 - 10^4 \text{ m}$
30 - 300 kHz	Low Frequencies (LF)	$10^4 - 10^3 \text{ m}$
300 kHz - 3 MHz	Medium frequencies (MF)	$10^3 - 10^2 \text{ m}$
3 - 30 MHz	High Frequencies (HF)	$10^2 - 10^1 \text{ m}$
30 - 300 MHz	Very high frequencies (VHF)	$10^1 - 1 \text{ m}$
300 MHz - 3 GHz	Ultra high frequencies (UHF)	$1 - 10^{-1} \text{ m}$
3 - 30 GHz	Super high frequencies (SHF)	$10^{-1} - 10^{-2} \text{ m}$
30 - 300 GHz	Extra high frequencies (EHF)	$10^{-2} - 10^{-3} \text{ m}$

- * Extremely low frequencies include power line frequencies
- * Frequencies of the range 300-3000 Hz are voice frequencies.
- * Very low frequencies are at the higher end of human hearing & these are used in defence communication.
- * Medium frequencies are used in AM radio broadcasting, marine & aerospace communication.
- * High frequencies are used in short waves radio broadcasting & military communication.
- * VHF & UHF are used in mobile radio, FM radio & TV channel communication.
- * SHF are known as microwave frequencies & are used in satellite & radar communication.
- * EHF are not yet used. Equipments required to generate & use these frequencies are complex & very costly.

⇒ Block Diagram representation of Communication System using modulation technique

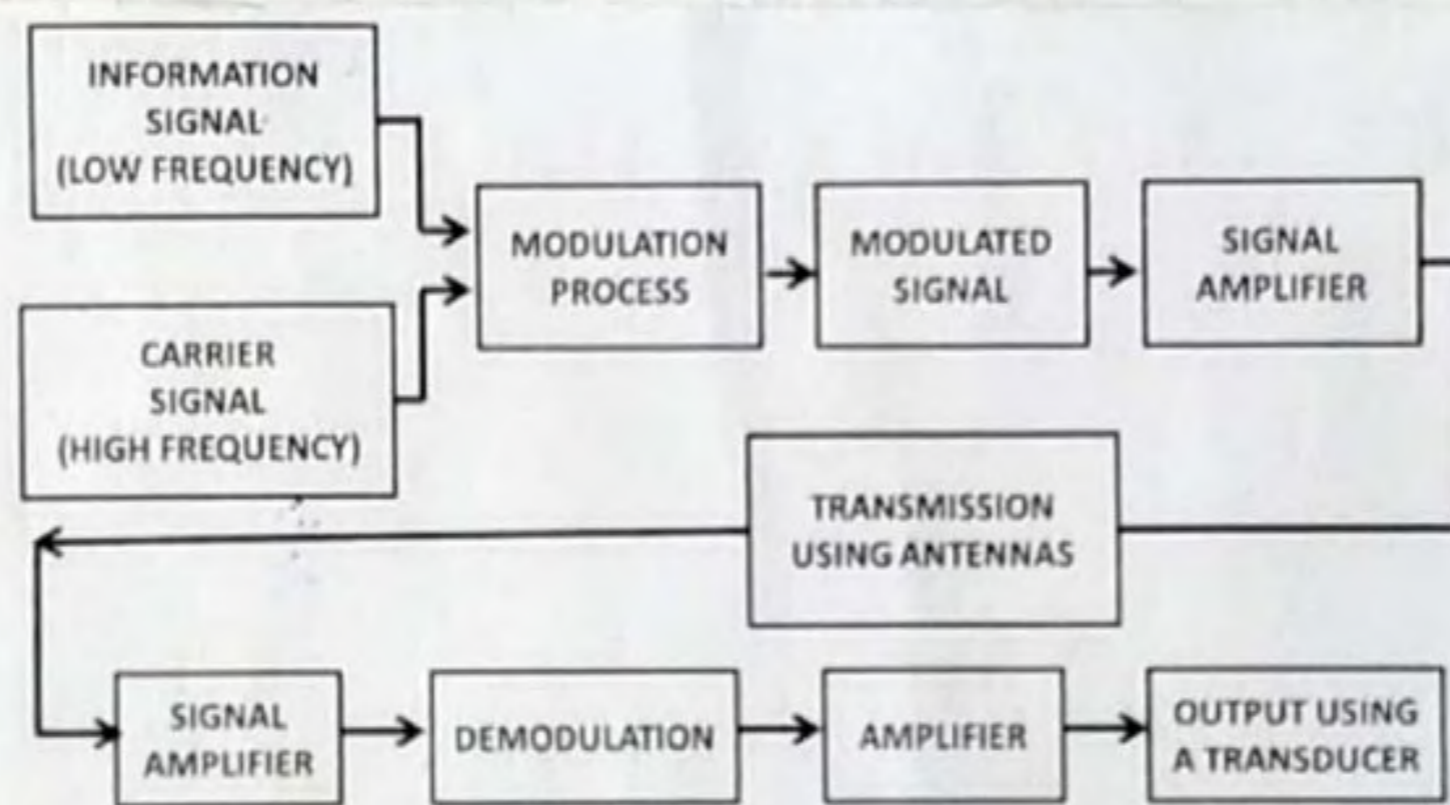


Fig: Block Diagram of Long Distance Communication System using Modulation Technique

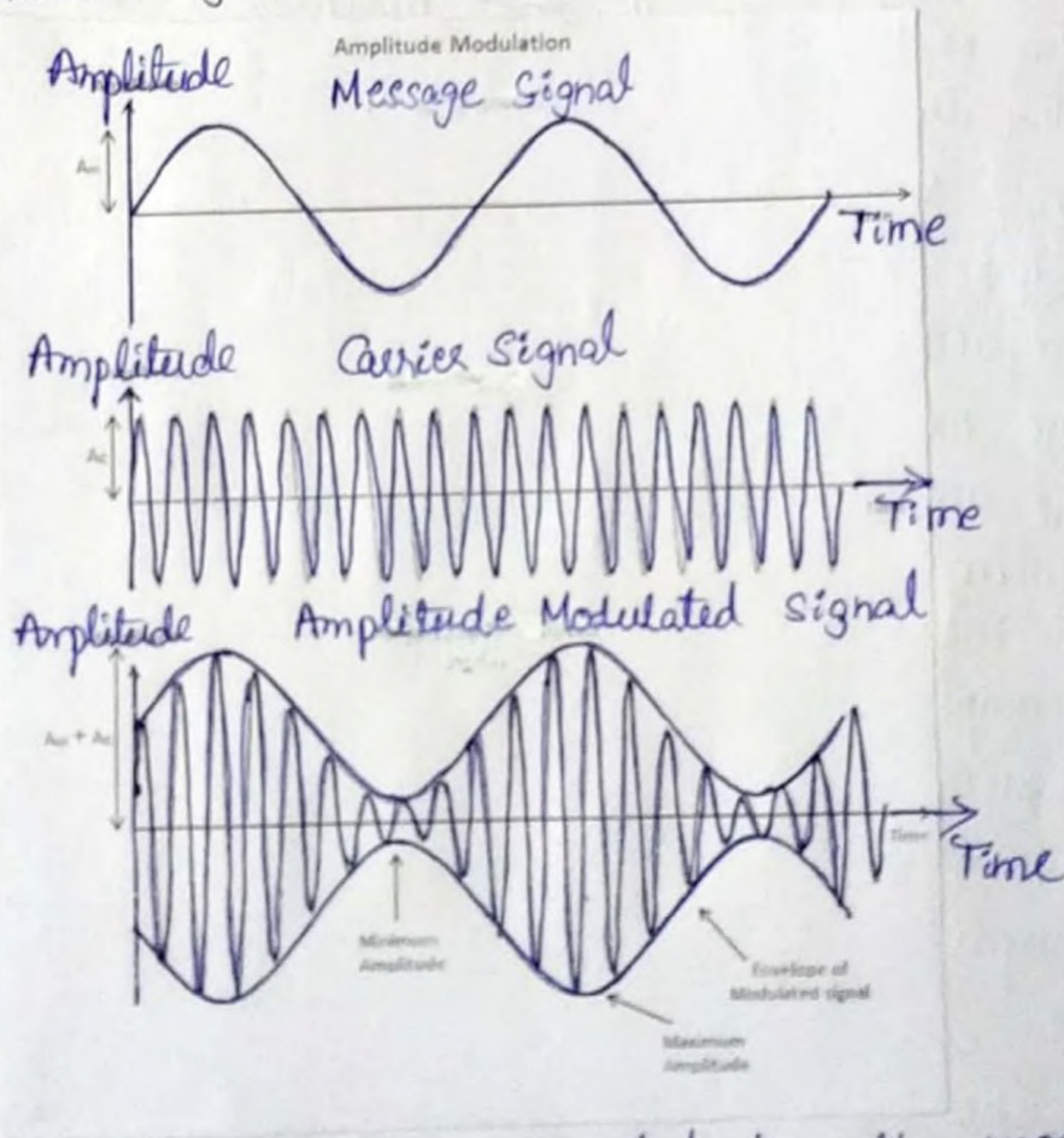
- * Information signal which is of low frequency cannot travel long distances.
- * Information signal is superimposed on the carrier signal which is of high frequency.
- * Using modulation process, the modulated signal that is produced is amplified using a signal amplifier.
- * The amplified signal is transmitted using antennas.
- * At the receiving end, amplification process takes place using a signal amplifier.
- * Using demodulation process, the original information signal is separated from the carrier signal & the information signal is again amplified & the output is given to the transducer.

⇒ Types of Modulation:

- * Information signals like sound signals are low frequency signals & as they have less energy they cannot travel long distance.
- * We can increase the energy by modulating them using high frequency carrier which increases the efficiency of transmission
- * Low frequency information signal gets superimposed on the high frequency carrier signal.

① Amplitude Modulation:

- * Definition: It is defined as the modulation in which the amplitude of the carrier signal is varied in proportion to that of the message signal.
- * The information signal is called the modulating signal.



* The carrier wave gets modulated by the modulating signal.

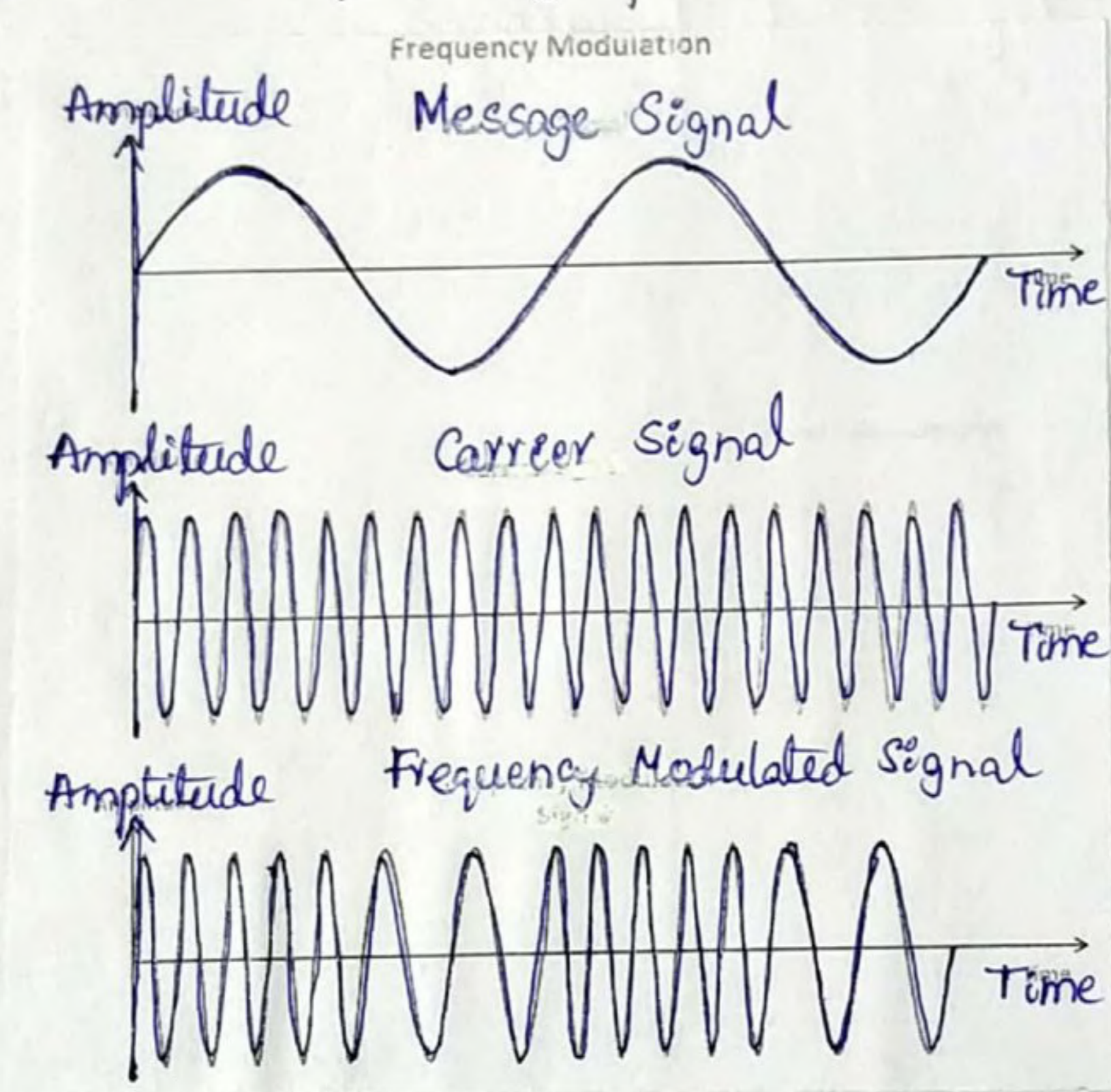
- ✗ A higher frequency signal containing the original information is created for long distance transmission.
- ✗ The frequency of the carrier signal remains the same but its amplitude varies in accordance with the modulating signal.
- ✗ An electronic circuit called a modulator is used for creating amplitude modulation.
- ✗ In amplitude modulation, the maximum value of the modulating signal V_m should be less than the maximum value of the carrier signal V_c .
- ✗ Modulation index is defined as the ratio of amplitude of modulating signal to the amplitude of carrier signal.
- ✗ The ratio of V_m and V_c is called modulation index.
- ✗ The value of modulation index should be less than 1. If V_m is higher than V_c , then modulation index becomes greater than 1 & such modulation creates distortion.

- ⇒ Frequency Modulation:
- ✗ Definition: It is defined as the modulation in which the frequency of the carrier signal is varied in proportion to the message signal.
 - ✗ As the amplitude of the information signal (i.e., the modulating signal) increases, the frequency of the carrier signal increases.
 - ✗ As the amplitude of the information signal decreases, the frequency of the carrier signal decreases.
 - ✗ Maximum frequency of the modulated signal occurs when the amplitude of the modulating signal reaches its maximum value.
 - ✗ When the modulating signal becomes negative, the frequency of the carrier will decrease and will continue to decrease till the modulating signal reaches its negative maximum value.

* Thus, frequency modulation creates compression and expansion of the modulated signal.

* The main advantage of frequency modulation over amplitude modulation is that a frequency modulated signal does not get affected by noise and it has higher transmitter efficiency.

* In frequency modulation, the modulating frequency creates a deviation in the frequency of the carrier signal.



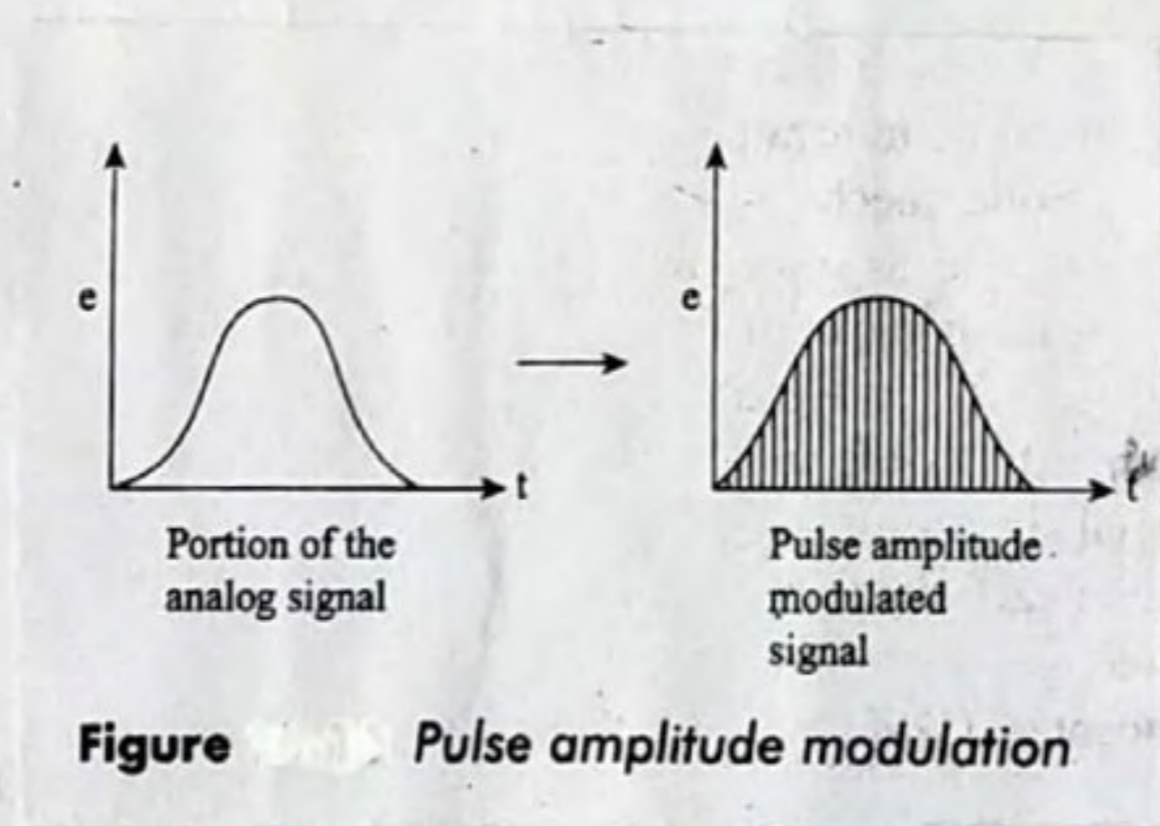
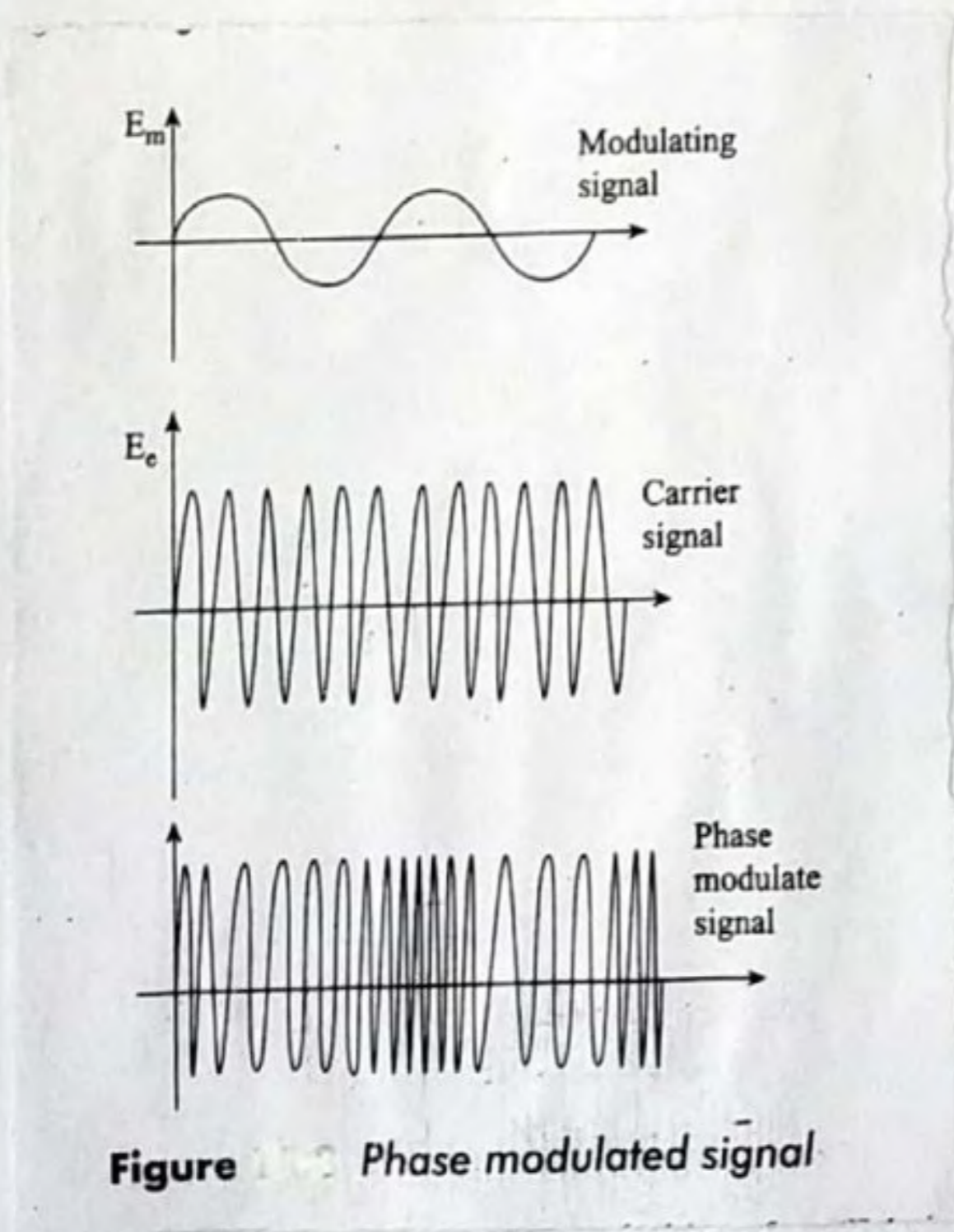
⇒ Phase Modulation:

Definition: It is defined as the modulation in which the phase of the carrier signal is varied in proportion to that of the message signal.

* In phase modulation, the phase deviation is proportional to the amplitude of the modulating signal and does not depend on its frequency.

- * Greater phase shift occurs if the amplitude of the modulating signal becomes greater.
- * Positive half of the modulating signal produces a lagging phase shift
- * Negative half of the modulating signal produces a leading phase shift.
- * When the amplitude of the modulating signal increases in the positive direction, the amount of phase lag increases.
- * As a result, the carrier frequency gets stretched out - it causes reduction of its frequency.
- * When the modulating signal increases in the negative direction, a phase lead is created, the carrier wave gets compressed & its frequency increases.

Comparison of AM and FM
 Analogies of AM



⇒ Comparison of AM and FM:

* The advantages of FM over AM are:

① Transmitting efficiency is higher

② Rejects interfering signals

③ Immune to external noise

* FM occupies more spectrum space because bandwidth of an FM signal is wider than that of an AM signal.

* The circuits needed to produce and demodulate an FM signal are more complicated than those required for AM signal.

⇒ Demodulation:

Definition: Demodulation is the process of recovering the original message signal from the modulated carrier signal.

* It is the process of extracting the original information from modulated carrier signal.

* After modulation, the modulated signal is transmitted in the form of electromagnetic waves using an antenna.

* A demodulator is used at the receiving end. It is an electronic circuit that is used to recover the information present in the modulated signal.

* Different kinds of circuits are used as demodulator depending on whether the carrier signal is amplitude modulated or frequency modulated or phase modulated.

* The process of demodulation involves 2 steps

1. Rectification & 2. Filtering

* The signal is first rectified & then filtered in the detector.

* The modulated wave is fed to the detector circuit for the purpose of demodulation.

* The diode rectifies the input signal. Negative portion of the modulated signal is removed.

- * The capacitor filter removes the high frequency carrier wave from the rectified wave.
- * The original signal is made available across the resistor.
- * Rectification is done before filtering otherwise the positive and negative envelopes of the modulated signal will cancel each other after filtering.
- * Quality of detector is measured in terms of linearity, sensitivity and signal handling capacity.
- * Linearity is measured as how accurately the o/p of the detector follows the input.
- * Sensitivity is the measure of how much of the i/p signal is delivered as the o/p.
- * Signal handling capacity is measured in terms of the ability of the detector in accepting a signal without any distortion.

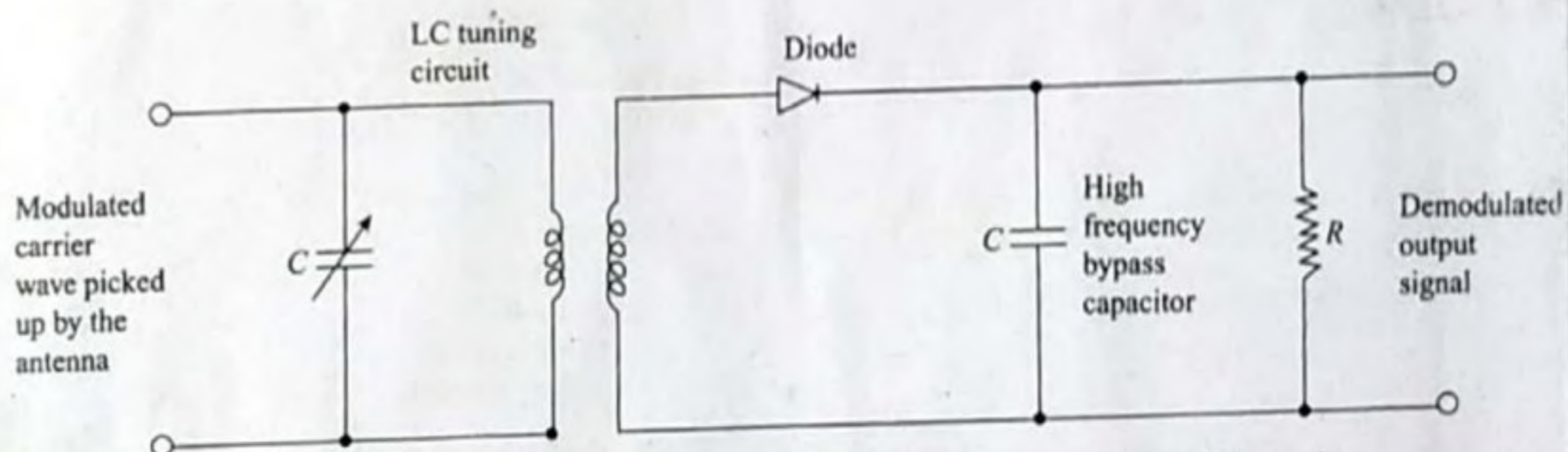


Figure 1.10 Basic circuit of a diode AM detector (demodulator)

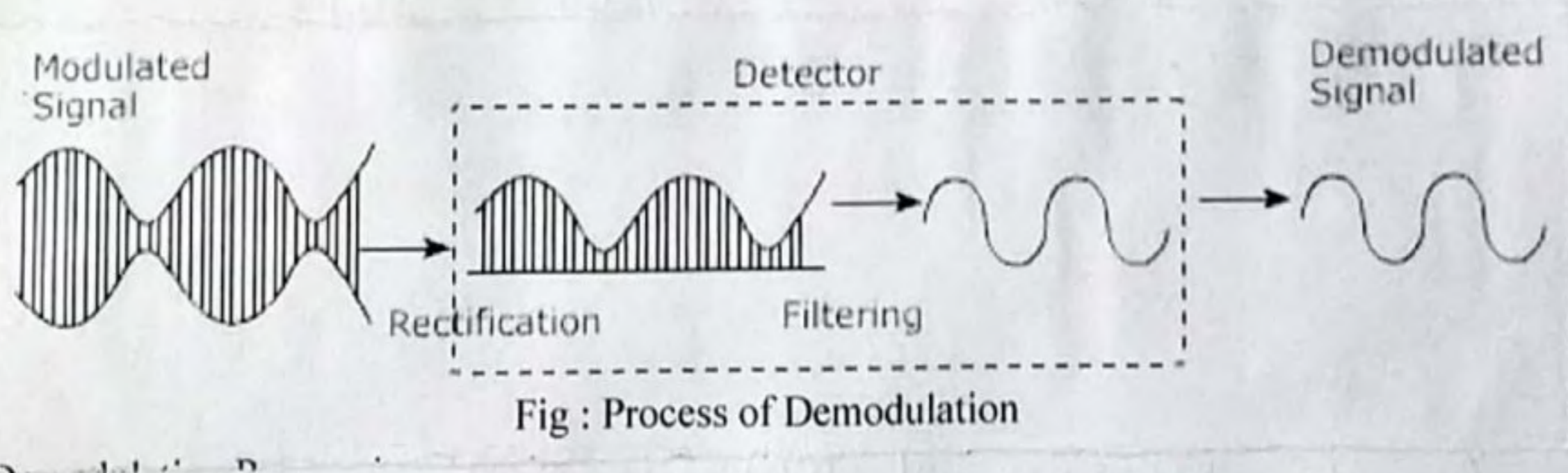


Fig : Process of Demodulation

⇒ Modulators:

Definition: Modulators are electronic circuits used to modify a carrier wave in accordance with the information signal to be transmitted.

Types of Modulators: ① Amplitude Modulator
② Frequency Modulator

⇒ Amplitude Modulator:

- * An amplitude modulation circuit will cause amplitude of the carrier vary in accordance with the information signal.
- * A high frequency carrier wave is generated by an LC oscillator circuit.
- * Carrier wave is passed through the amplifier circuit whose gain can be varied in accordance with the modulating signal.
- * The circuit contains a mixing network of resistors, a diode rectifier and LC tuning circuit.
- * The carrier wave generated by the oscillator is connected to resistor R_1 and the modulating message signal from the transducer is connected to resistor R_2 .
- * A transducer converts the information signal into an electrical signal.
- * The network containing resistors R_1 and R_2 adds the two signals linearly. One signal is the information signal and the other signal is the carrier signal.

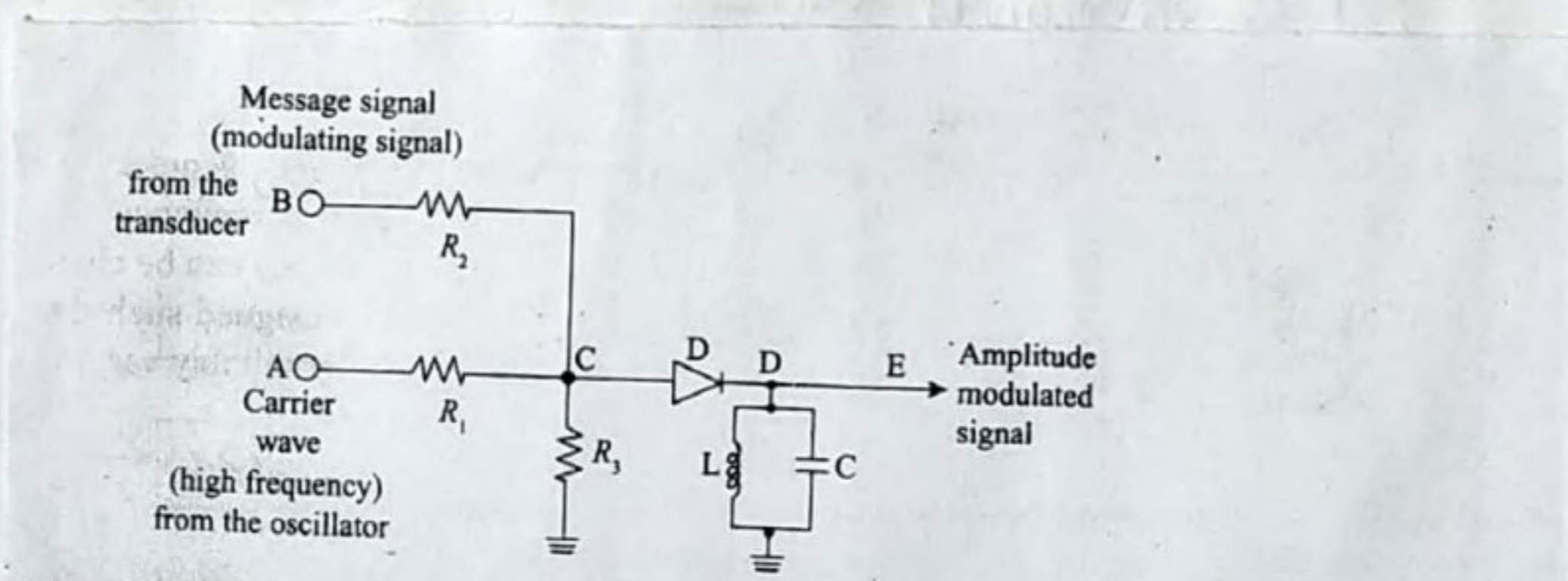


Figure Amplitude modulation circuit in its simplest form

→ If the two signals are sine waves of different frequencies, their sum is shown in figure "

→ Diode D cuts off the negative portion of the input signals.

→ The positive pulses are applied to the tuned circuit made of L and C.

→ The output of the circuit is an amplitude modulated signal.

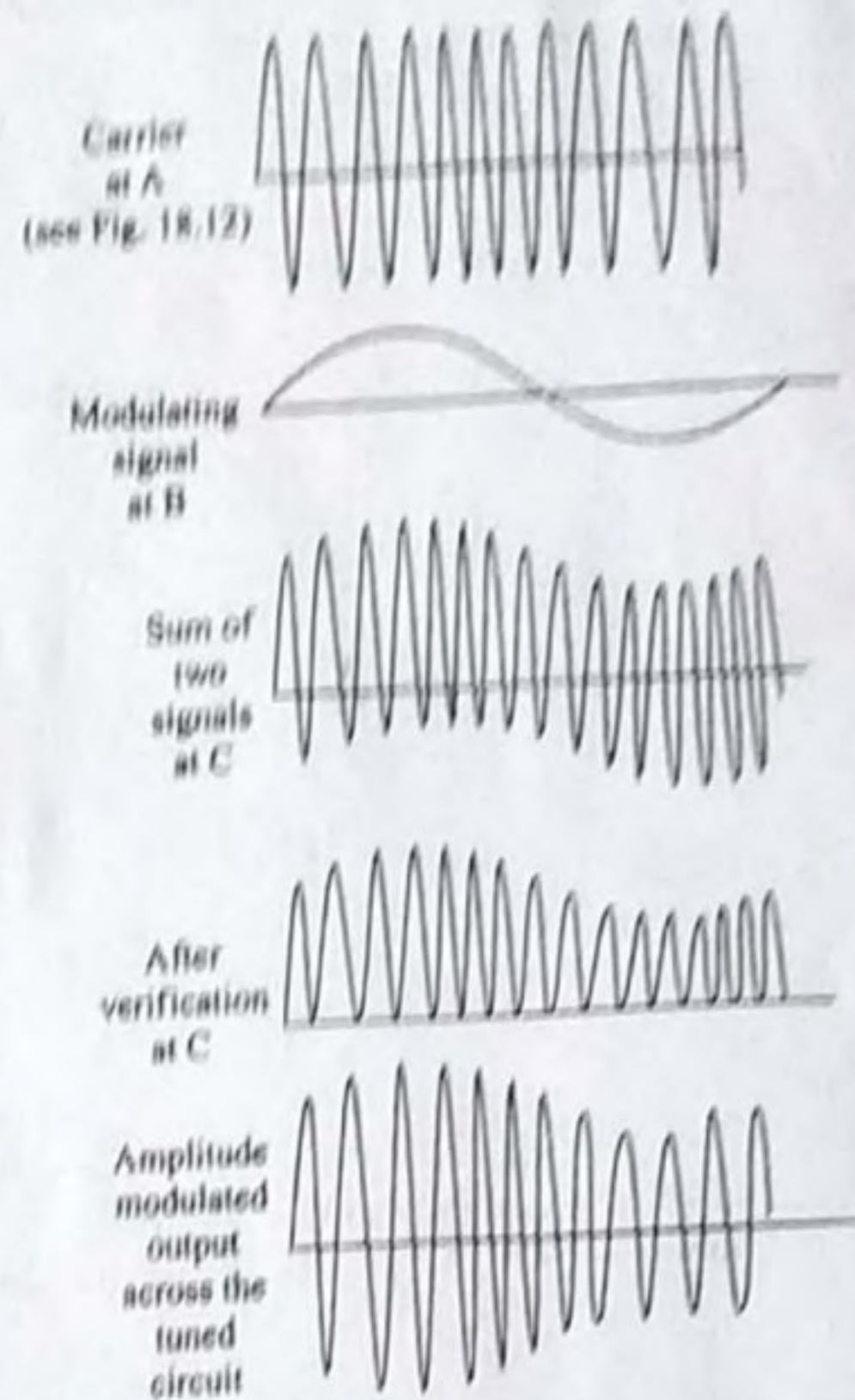


Figure 18.12 Amplitude modulator wave forms

⇒ Frequency Modulator:

* A frequency modulator is an electronic circuit which causes frequency of the carrier to vary in accordance with the information signal.

* Carrier frequency is generated using an LC circuit.

* In an LC oscillator, the frequency of the carrier wave produced is determined by the values of L and C.

* The carrier frequency can be changed by changing the values of inductance and capacitance.

⇒ Multiplexing:

Definition: Multiplexing is the process by which a large number of information signals can be communicated through a single communication channel.

* The communication channel could be a cable connection or a radio link.

* Advantages of multiplexing are:

- ① Less complex
- ② Inexpensive
- ③ Cost reduction
- ④ Conservation of spectrum space.

* Multiplexing is used in

① Telephone Communication

③ Telemetry

② TV stereo broadcasting

④ Radio broadcasting

* Types of Multiplexing:

① Frequency division multiplexing (FDM)

② Time division multiplexing (TDM)

⇒ Frequency Division Multiplexing (FDM):

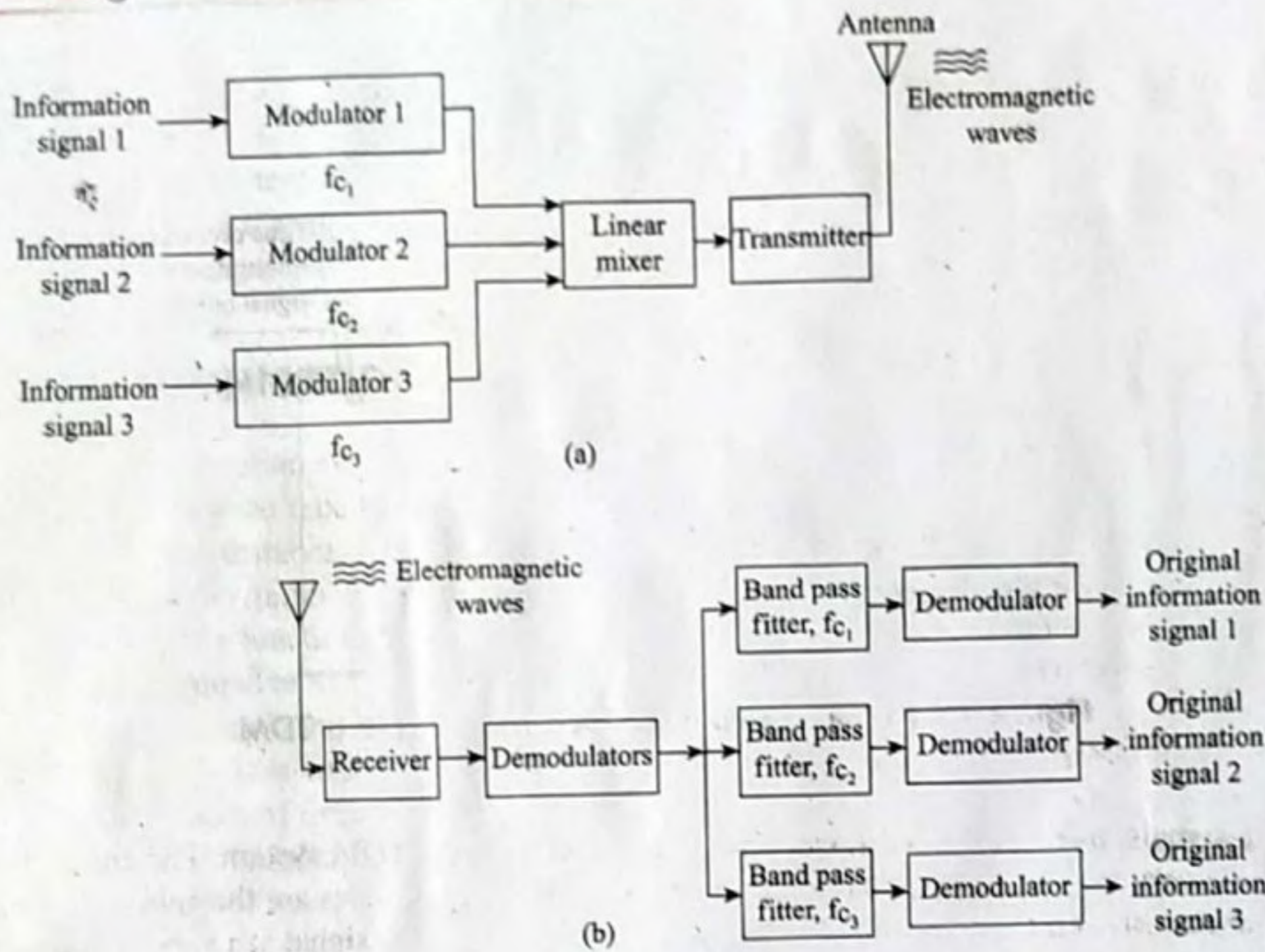


Figure Schematic representation of frequency division multiplexing: (a) transmitting end; (b) receiving end

* In FDM, multiple signals are adjusted within the bandwidth of one communication channel.

* Bandwidth is the range of frequencies over which a communication channel is established.

* Each signal occupies only a small portion of the available bandwidth and the signals do not get mixed up.

* In FDM, different signals are assigned different frequencies within the given bandwidth.

- * Information signals are given to the modulators.
- * The modulated carrier signals are mixed i.e., summed up in a linear mixer.
- * All the carriers are combined into a composite signal & this composite signal is transmitted using a transmitter over a single communication channel using an antenna.
- * At the receiving end, an antenna receives the electromagnetic signal and a demodulator demodulates it and gives out the composite signal.
- * Using bandpass filters and demodulators each input information signal is recovered.
- * The band pass filter used at the receiving end passes only its channel and rejects others.
- * In FDM large number of signals are transmitted using a single channel sharing the channel bandwidth range.
- * This is achieved by assigning each information signal a small portion of spectrum space within the channel bandwidth.

Applications of FDM:

- ① Telemetry
- ② Telephone Systems
- ③ Stereo FM

⇒ Time Division Multiplexing (TDM):

- * In TDM, multiple signals are transmitted using different time slots.
- * Using time sharing mechanism multiple signals are transmitted over a single channel.
- * The whole bandwidth that is available can be utilized by each signal on time sharing basis.

- * Among the multiple signals that are transmitted, each signal is allowed to use the channel for a short period of time turn by turn.
- * A mechanical switching arrangement is made using electronic switching circuits.
- * Transmission of all the signals is done continuously by time sharing.
- * Time division multiplexing is possible for both analog and digital signals.
- * Equal time slot is allocated for each signal and a particular sequence is maintained.
- * Ex: If 4 digital signals are to be transmitted over a single channel, the signals will take their turn as 1 2 3 4 1 2 3 4 1 2... This is the concept of TDM.
- * Analog signals which are transmitted using TDM system are sampled at a very high rate using electronic switching technique.
- * Samples are the values of the analog signal after short interval of time.
- * By taking samples of the analog signal at a very high rate, pulses are produced whose magnitude vary in accordance with the analog signal.
- * This process of sampling an analog signal is called pulse amplitude modulation (PAM).

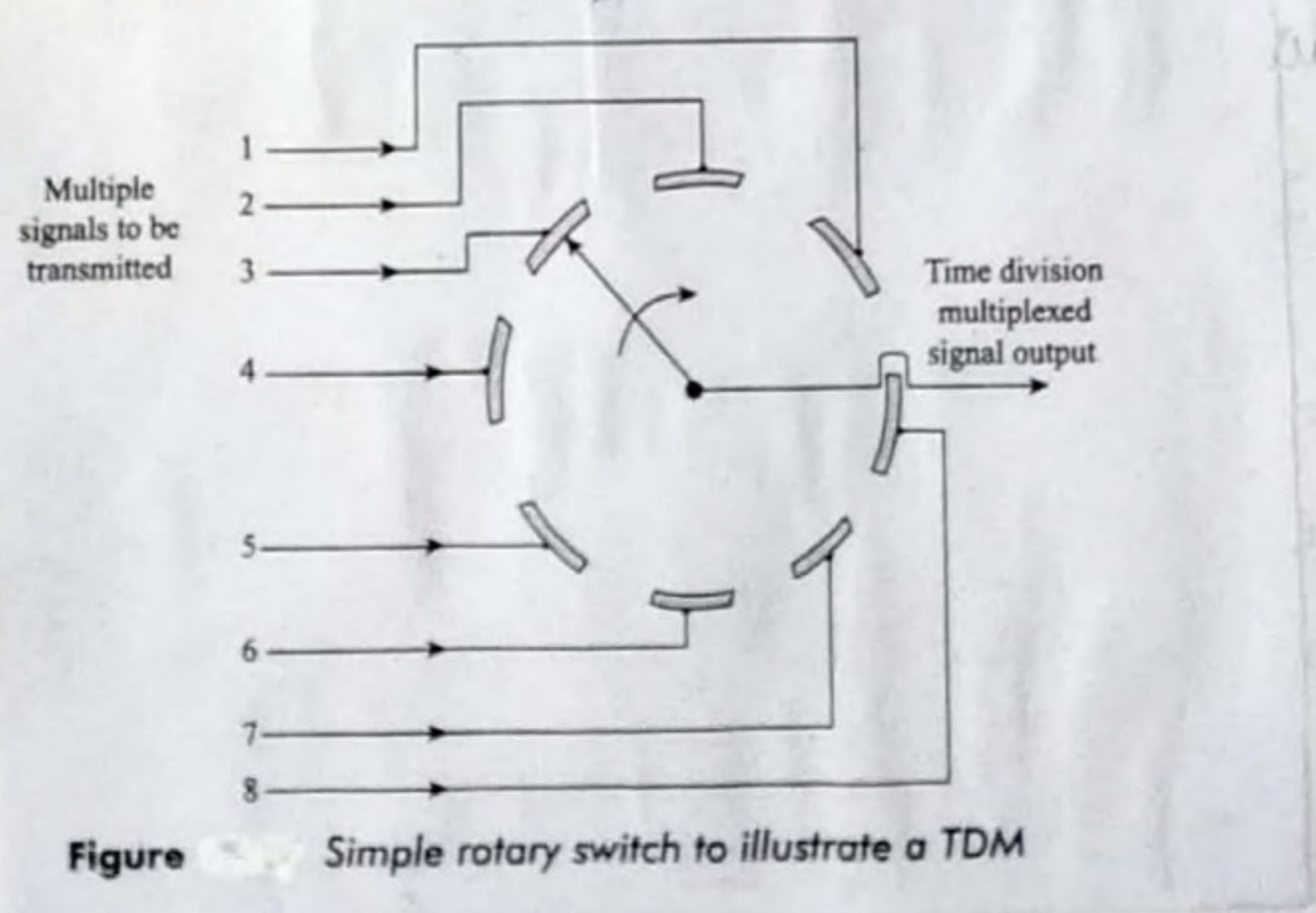


Figure Simple rotary switch to illustrate a TDM

⇒ Communication Receivers:

- * In wireless communication system, the transmitted signal becomes very weak when it reaches the receiver as it travels long distance & may get distorted due to noise or unwanted signals.
- * Function: The function of receiver is to receive only the desired signal and reject all other signals.
- * A simple receiver is a tuned circuit. There are LC circuits tuned to resonate at a desired signal frequency.
- * If the resonant frequency matches with the frequency of the signal to be received, then the circuit will pass that signal & reject other unwanted signals.
- * Selectivity: The selectivity of a receiver is the ability to select a desired frequency signal & rejecting the unwanted noise generating signals.
- * Sensitivity: The sensitivity of a receiver is the ability to pick up weak signals. Those signals must be amplified using an amplifier and then demodulated.

⇒ Functions of Radio Receiver:

- ① Select the desired signal from all other unwanted signals
- ② Amplify the desired signal.
- ③ Demodulate the amplified signal i.e., separate the message signal from the carrier signal.
- ④ After demodulation, amplify the original message signal.
- ⑤ Apply the amplified demodulated signal to the transducer

⇒ Types of Radio Receivers:

- ① Tuned Radio Frequency Receiver
- ② Super Heterodyne Receiver.

⇒ Tuned Radio Frequency Receiver:

- * The TRF receiver block diagram consists of the following blocks.
 - ① RF Amplifier
 - ② Detector / Demodulator
 - ③ Audio Amplifier
 - ④ Power Amplifier

- * Electromagnetic waves are received by the antenna.
- * Two or three stages of radio frequency amplification is done using RF amplifiers between the antenna & the detector.
- * The RF amplifiers amplify the amplitude of the received signal.
- * To improve selectivity and sensitivity, many amplifiers are used.
- * The signal received should be strong enough to produce strong output.
- * Radio frequency receivers are tuned to select and amplify the desired signal.
- * The amplified signal is detected by a diode detector and demodulated using a bypass capacitor.
- * Due to demodulation, the carrier signal is bypassed and the original information signal is recovered.
- * This signal is amplified to the necessary power level using audio & power amplifiers before connecting to the speaker.
- * TRF receivers have some problems with respect to selectivity & variation of bandwidth over the tuning range. These problems were overcome in the improved type of receiver called superheterodyne receiver.

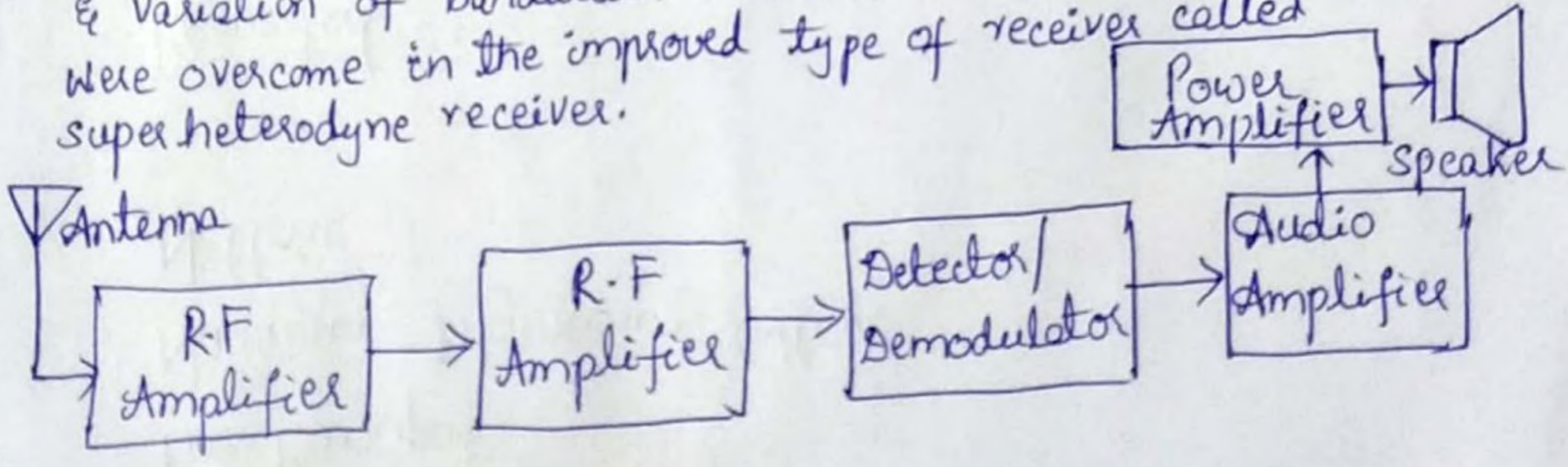


fig: Block diagram of TRF receiver

⇒ Super heterodyne Receiver:

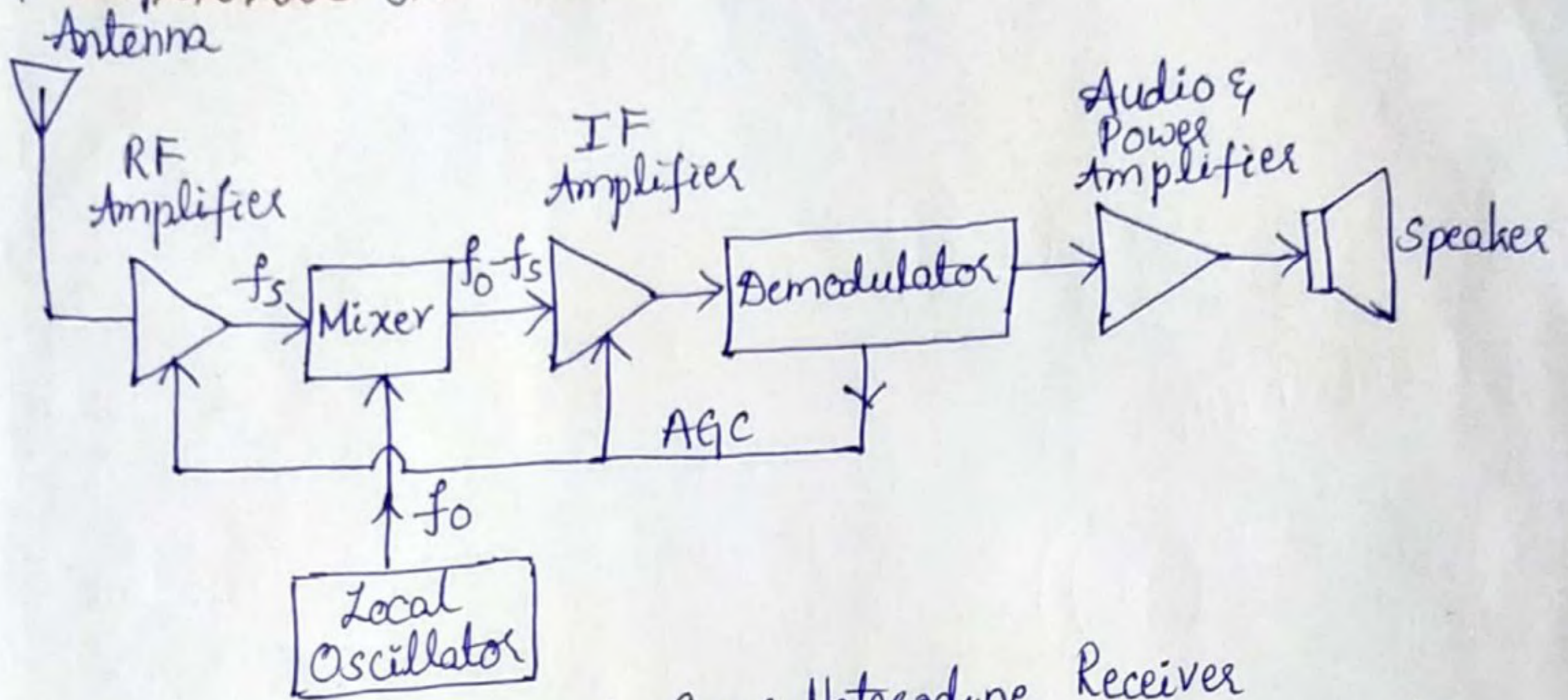


fig: Block diagram of Super-Heterodyne Receiver

- * A super heterodyne receiver uses frequency mixing technique to convert all incoming signals into a fixed lower intermediate frequency which can be more conveniently processed than the radio frequency.
- * The antenna picks up the radio signal & feeds it to the RF Amplifier.
- * Amplification of the signal is needed at the ip stage as the received signal is very weak.
- * A tuned circuit along with RF amplifier improves sensitivity due to extra gain & improves selectivity.
- * The local oscillator produces sine wave signals.
- * The local oscillator is tunable i.e, its frequency can be adjusted.
- * By tuning the local oscillator frequency, the ip frequency ^{at the mixer} can be converted to an intermediate lower frequency. This is called Superheterodyning.
- * If f_s is the frequency of the received signal
 f_o is the frequency of the signal generated by the local oscillator
 the o/p of mixer will be $f_o - f_s$. This is called intermediate frequency (or) IF.

* The original audio signal containing IF is amplified by the IF amplifier.

* The original audio signal is recovered using a demodulator.

* The recovered signal is further amplified & applied to the loud speaker.

* The Automatic Gain Control (AGC) controls the gain of RF & IF amplifiers to maintain constant o/p if the strength of the received signal at the input stage is fluctuating.

⇒ Microwave Communication:

* The frequency spectrum used for radio communication has become overcrowded due to large number of users.

* Increased number of users have crowded the airspace. Hence radio communication has been extended to higher frequencies.

* Microwaves are signals with frequencies higher than 1 GHz.

* In electronic communication system, the modulated signal along with its sidebands occupies certain space in the radio frequency spectrum. The small portion of spectrum is called channel.

* Spectrum space is becoming congested due to introduction of large number of communication channels.

* To solve this problem, higher frequencies are used.

* Microwave region has sufficient space for new channels and of wider bandwidths.

* Microwave frequency spectrum is used for long-distance telephone communication.

* Radio detection & ranging (Radar) communication uses microwave frequency.

* Radar is used in detecting the presence of distant object like an aeroplane or ship.

* The radar system transmits high frequency signal which gets reflected from the object to be detected.

* TV stations use microwave to transmit TV signals.

* Microwaves are used in satellites in space communication.

* Microwaves are used in microwave ovens in kitchens, hospitals for heating purpose.

⇒ Satellite Communication:

* Artificial satellites are launched to go into orbit around the earth.

* Satellite is launched with both vertical & forward motion.

* The solar panels attached to the satellite supply power to the satellite.

* The action of transmission of signal from the ground station to the satellite is called uplink.

* The action of receiving signals from the satellite to the ground station is called downlink.

* The name for the transmitter-receiver combination in a satellite is called transponder.

* The role of transponder is to amplify the signal before it is transmitted back to the earth station.

* Applications:
Used in television & mobile communications
Used in surveillance or observation
Used in Geological Information System (GIS)

* Advantages: It is meant for long distance communication.
Reliability & quality of communication is very high.
It is possible to have a very wide transmission-reception area using a satellite.
Communication link can be established to very distant remote areas.

* Block diagram of Satellite Transponder:

The satellite transponder receives uplink transmission information & retransmits the signal through its downlink mechanism.

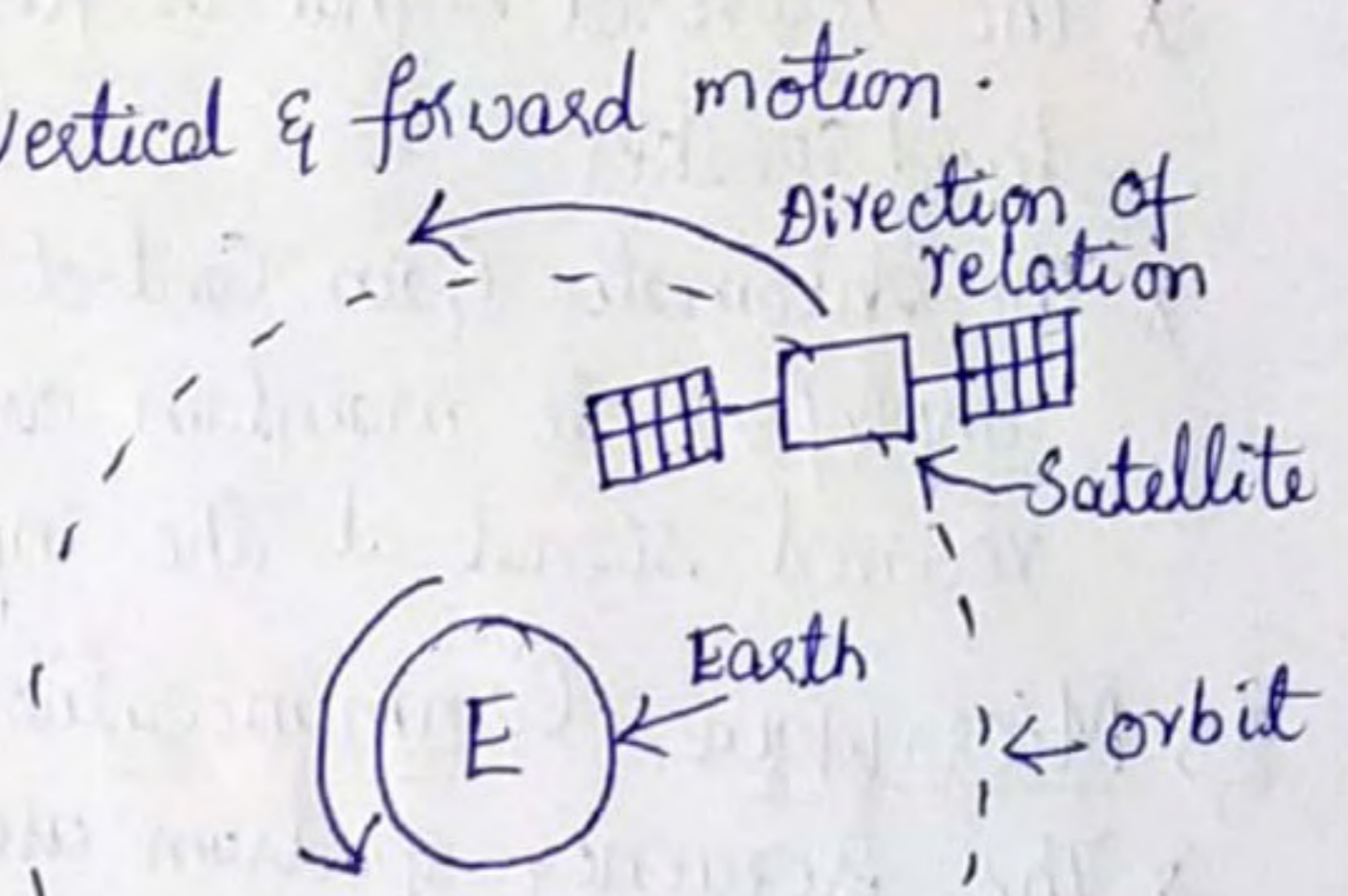


fig: Artificial satellite orbiting round the earth

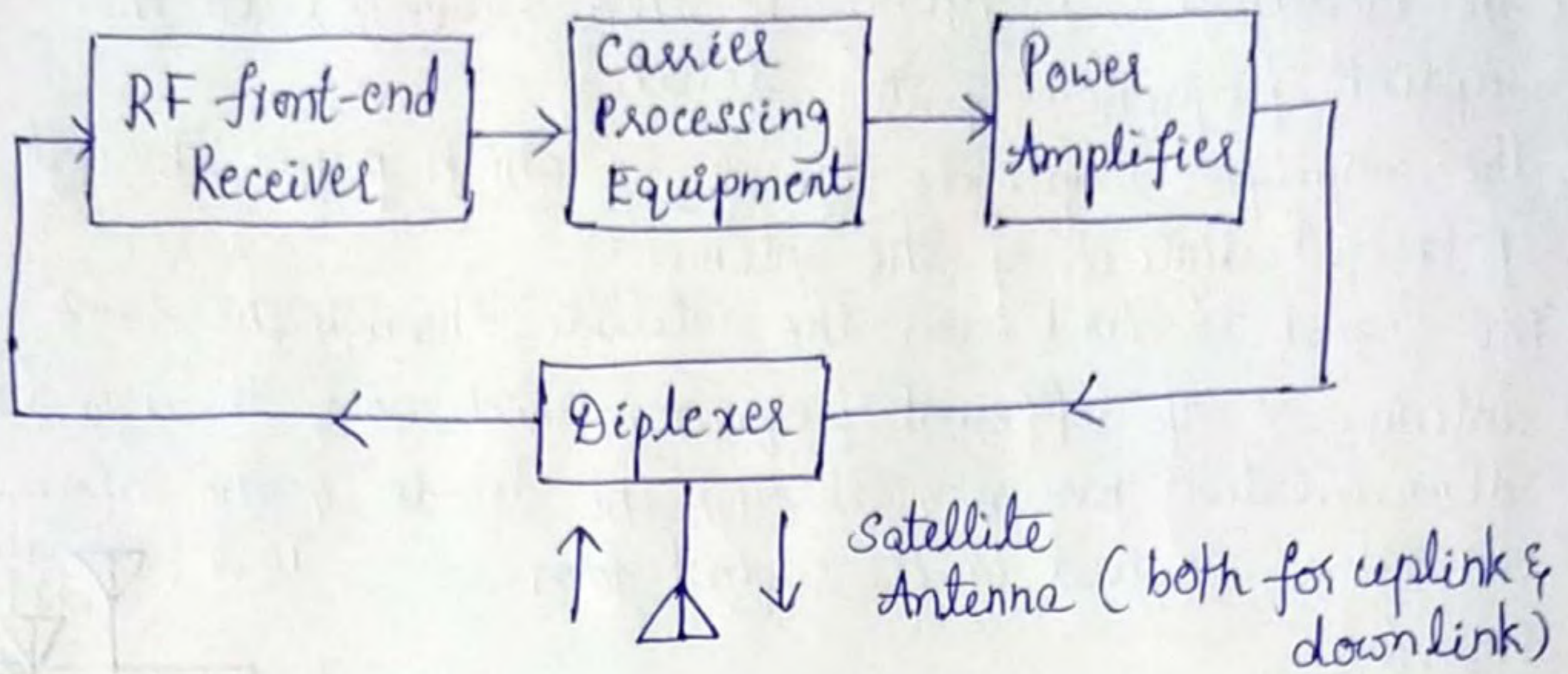


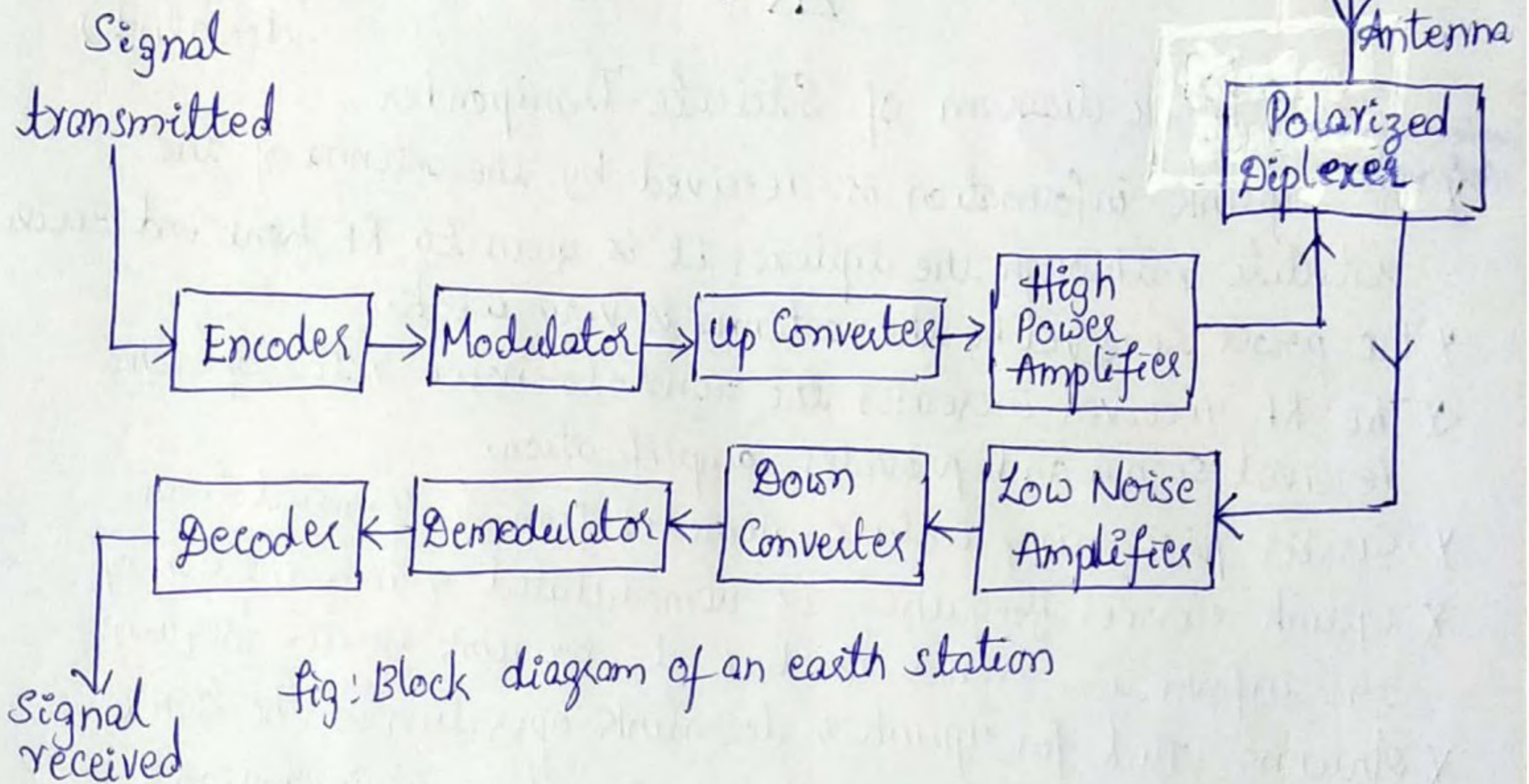
Fig: Block diagram of Satellite Transponder

- * The uplink information is received by the antenna of the satellite & through the diplexer it is given to RF front end receiver.
- * The power received by the antenna is very weak.
- * The RF receiver increases the signal-to-noise ratio of the received signal and provides amplification.
- * Carrier processing includes demodulation & remodulation.
- * Uplink carrier frequency is demodulated & remodulation of the information signal is done at downlink carrier frequency.
- * Antenna used for uplink & downlink operation is the same.
- * As two frequencies are different, simultaneous reception & transmission is possible using a diplexer through single antenna.
- * Power amplifier is used to increase the power level of the transmission signal so that information reaches the earth station.

⇒ Block Diagram of Earth Station:

- * The encoder converts the baseband information signal into a format ready for modulation.
- * The carrier is modulated by the encoded information signal.

- * The modulated frequency is then converted to the uplink frequency of the satellite.
- * The amplifier amplifies the signal which passes through polarized diplexer of the antenna.
- * The signal received from the satellite through the same antenna is of different frequency which goes through the demodulation process & through the decoder & the information signal is received in its original form.



⇒ Fibre-Optic Communication:

- * Communication can also be made using a light wave.
- * Information can be put on a light beam & transmitted through fibre-optic cables.
- * Fibre-optic cable is a narrow light pipe that is used to carry light beam from one end to the other.
- * If the information signal is a voice signal, it is converted into digital signal using A to D converter.
- * A light source (LED) is used which is made on and off by the digital data.

- * The light beam pulses are created by the digital information signal are fed to the fibre-optic cable for long distance communication.
- * The light source entering the optical fibre cable comes out at the other end of the cable.
- * The opt light pulse is much less in amplitude than the input due to attenuation i.e., loss of light energy as the light pulse travels through the fibre cable.
- * The loss of light energy is due to absorption, scattering & dispersion in the core of the fibre cable.
- * Fibre optic cable is a long thin strand made of glass or plastic.
- * The opt of the fibre optic system is a detector which will sense the light source & converts it into an electrical signal.
- * The signal is then amplified & transformed into original digital signal or data.
- * At the receiver a photodiode is used which converts light into voltage pulses.
- * For the communication to be very fast, the speed at which the optical pulses are created must be very high. This is called pulse rate.

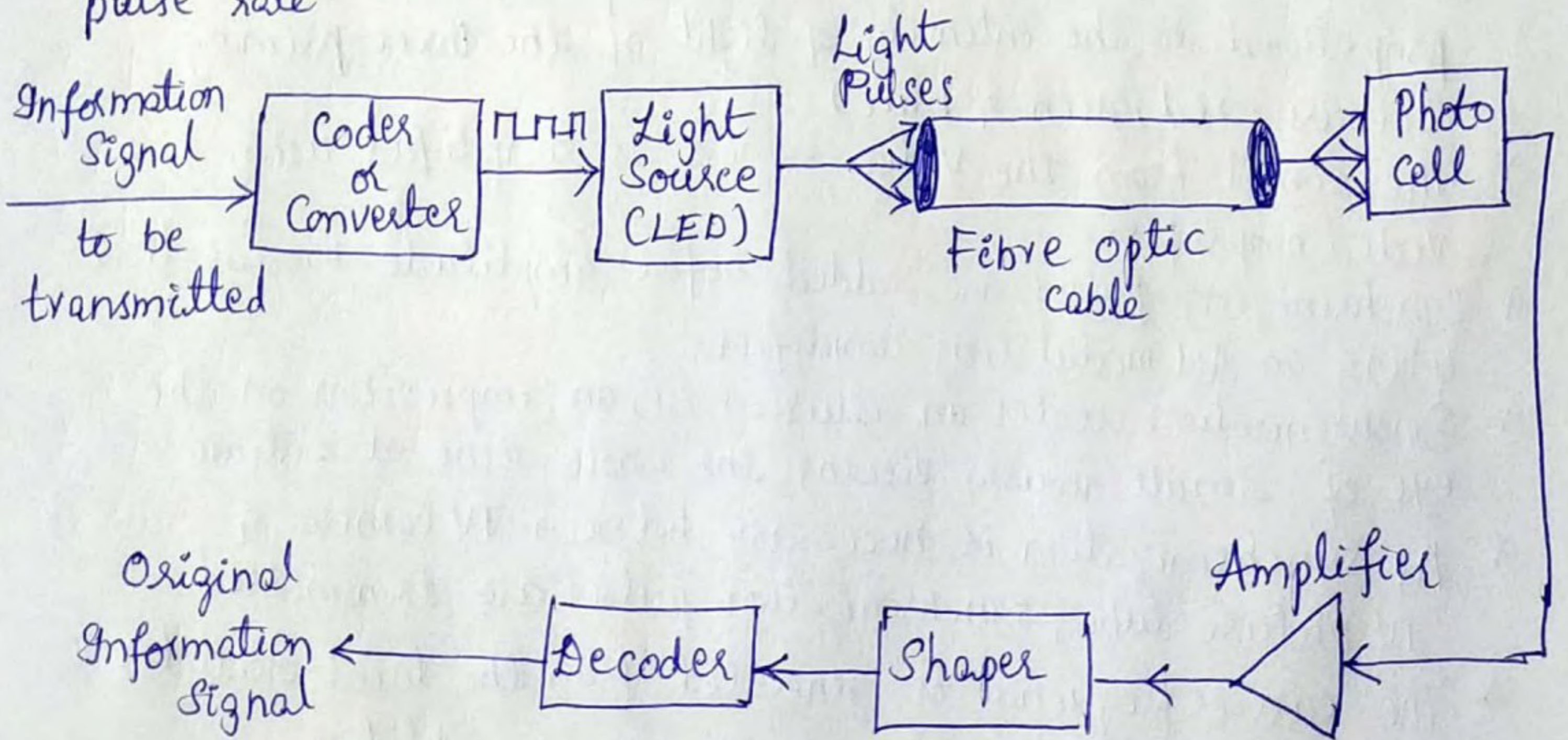


fig: Block diagram of fibre-optic Communication System

⇒ Television:

- A television system transmits a picture or a scene along with sound to a distant place.
- * The picture signal is amplitude modulated & the sound signal is frequency modulated.
 - * The two carrier frequencies are separated by a wide margin to avoid mixing of the two.

⇒ TV Transmitter:

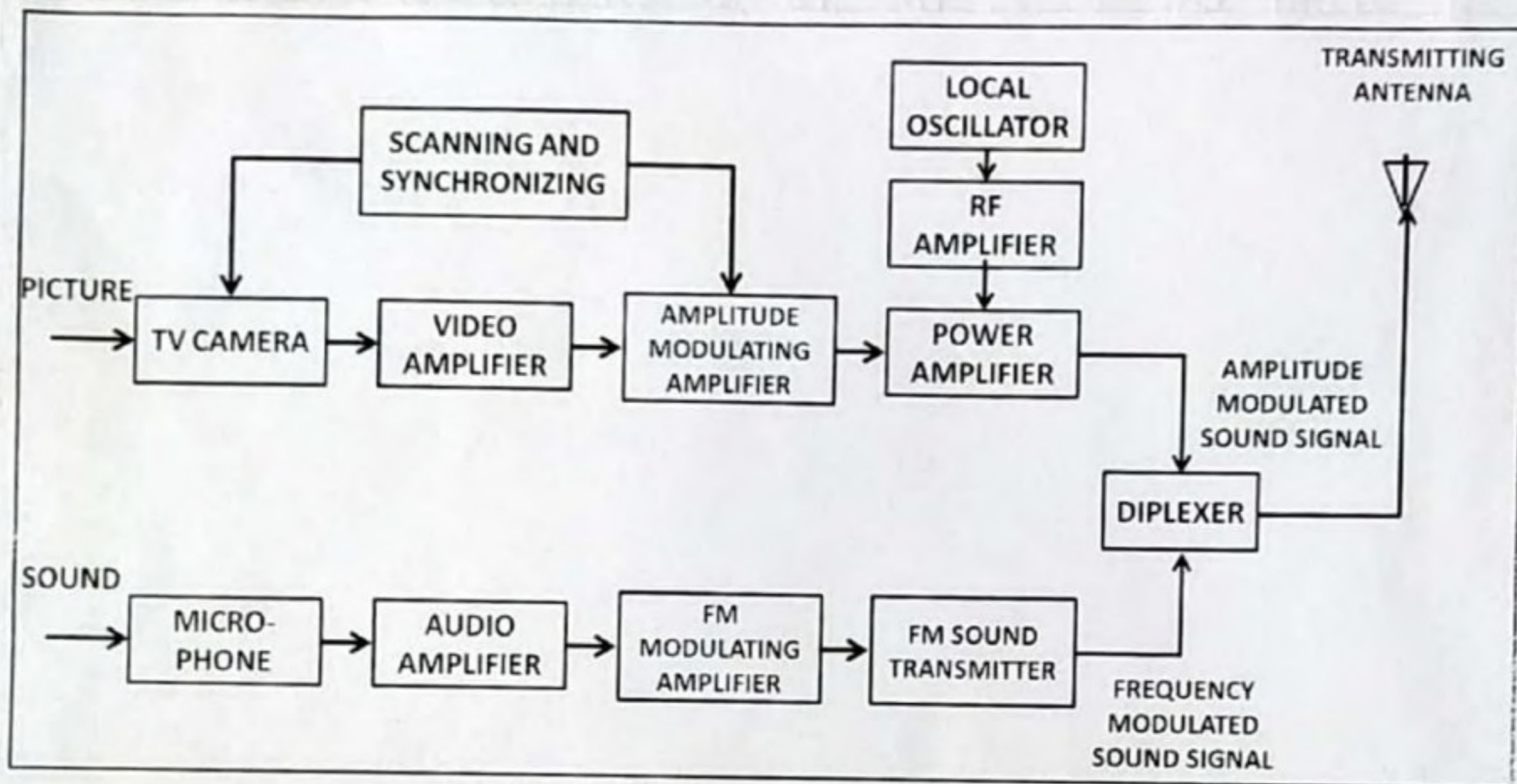


Fig: Block Diagram of TV Transmitter

- * A color TV camera is used to develop signal voltages proportional to the intensity of light of the three primary colours (red, green & blue).
- * The signal from the video camera is amplified using a video amplifier.
- * Synchronizing pulses are added before amplitude modulation using an AM modulating amplifier.
- * Synchronization creates an illusion i.e., an impression on the eye of simultaneously viewing the whole scene at a time.
- * As synchronization is necessary between TV camera & TV picture tube, synchronization pulses are transmitted.
- * The carrier frequency is generated from the local oscillator which is amplified & fed to the power amplifier.

- * The carrier is modulated by the modulating signal received from the modulating amplifier.
- * This signal is mixed with the audio signal & transmitted through the antenna.
- * The sound signal is picked up using a microphone which is amplified & then frequency modulated using a carrier frequency.
- * Frequency modulated sound signal is combined with amplitude modulated picture signal before transmission.

⇒ TV Receiver:

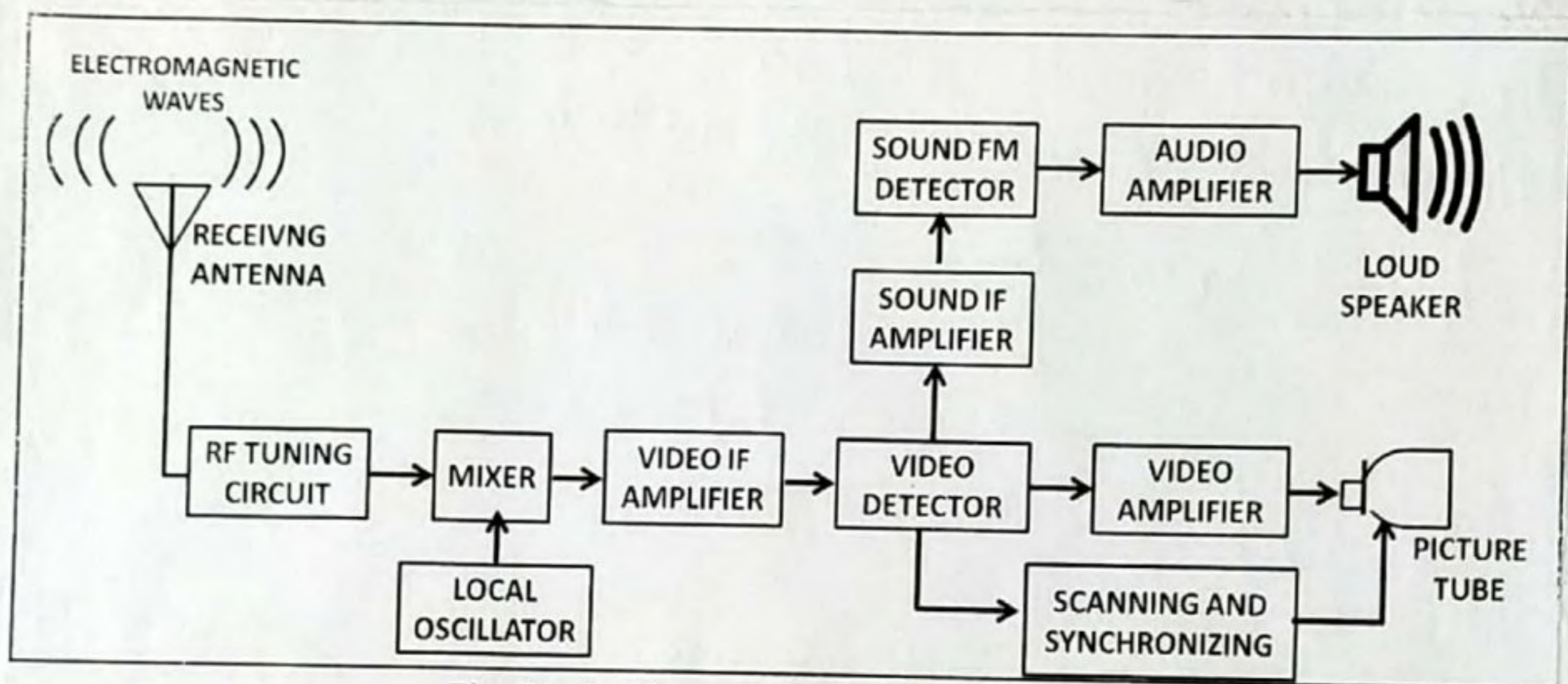


Fig: Block Diagram of TV Receiver

- * TV transmitter antenna radiates electromagnetic waves.
- * At the receiving end, a receiving antenna intercepts the radiated RF signals (radio frequency signals).
- * An RF tuner circuit selects the desired channel signals of a particular frequency band & converts into IF (intermediate frequency) band of frequencies using a local oscillator.
- * Video IF amplifier amplifies the signal.
- * Video detector isolates video & audio signals (separates 2 signals).
- * The video signals are amplified using video amplifiers & are sent to picture tube.
- * In the picture tube the electrical signal is converted to picture on its screen.

* The frequency modulated sound signal is demodulated after amplification.

* The audio o/p is connected to the loudspeaker where the electrical signal is converted to sound.

⇒ Need for Synchronization:

* When a TV camera starts scanning line 1, the receiver i.e., picture tube should simultaneously start scanning line 1.

* The speed of scanning of the video camera should be exactly the same as the speed of scanning process of the receiver.

* Hence synchronization is essential to avoid distortion of the picture received on the TV screen.

⇒ Mobile Communication:

* A mobile telephone is a wireless telephone capable of operating while moving & can be carried by a person.

* Cellular telephone systems offer high user capacity within limited frequency spectrum.

* It also allows large number of users to share limited number of radio channels.

* Each radio coverage area is divided into hexagon-shaped cells.



Fig: Honey Comb Cell Pattern



Fig: Transmitter located at the center of the cell

* A geographical area is divided into small cells.

The design process involves allocation of channel groups to the cell

* The size of the cell area changes depending upon user density & calling pattern.

* Transmitter is located at the centre with omnidirectional antenna.

* An omnidirectional antenna radiates & receives signals equally in all directions.

* A base station located at each cell serves as a control station for all users in that cell.

⇒ Cellular Telephone Network Components:

* The six essential network components are:

① Electronic Switching Centre:

* This is the digital telephone exchange located at the

MTSO (Mobile Telephone Switching Office).

* It controls switching between mobile to mobile, mobile to wireline & wireline to mobile communications.

* The main function of MTSO is to control assigning of channels, call setup & call termination which includes signalling, switching, supervision & allocation of channels.

② Base Station (or) Cellsite Controller:

* It turns the radio transmitter & receiver on & off, manages each of the radio channels at each cellsite & performs tests on the cell-site equipment.

③ Radio Transmitter & Receiver:

* Each cellsite contains one radio transmitter & two radio receivers tuned to the same channel frequency.

* Transceiver is a transmitter-receiver combination which has a common antenna to transmit & receive signals.

* In CDMA, one unique code is assigned to each subscriber. This code is used to encode & decode information signal.

① Space Division Multiple Access (SDMA):

SDMA technique involves control of radiated energy from the antennas for each subscriber by using directional antennas at each cell site.

* The techniques involves in multiple access methods are

FDMA: dividing the spectrum into different frequency channels

TDMA: time sharing of the spectrum by the subscribers

CDMA: Using the same spectrum simultaneously by all the subscribers by the use of coding.

SDMA: sharing of the spectrum in angular directions by using smart antennas.