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15.2.2 General Gas Law:

In order to derive general gas law, we make use of Boyle's law, Charles's law and Avogadro's law. An interrelation among the physical quantities e.g. pressure, volume, temperature and amount of matter of a given sample of gas is termed as "equation of state" of gas or General gas law.

DO YOU KNOW?
Real or permanent gas is a gas that obeys gas Laws at high temperature and low Pressure only.

According to Boyle's law:

$V \propto \frac{1}{P}$ (when n number of mole and temperature T are kept constant)

According to Charles's law:

$V \propto T$ (when n and pressure P are kept constant)

According to Avogadro's law:

$V \propto n$ (when T and P are kept constant)

DO YOU KNOW?
1dm = 0.1m
Hence $1dm^3 = 10^{-3}m^3$
Ideal or perfect gas is a gas that obeys gas Laws at all temperatures and pressures.

Consider for a moment that none of the variable are to be kept constant, then all the above three relationships can be joined together.

$$V \propto \frac{nT}{P}$$

$$V = constant \times \frac{nT}{P}$$

$$V = \frac{RnT}{P} \quad \ldots\ldots\ldots\ldots\ldots\ldots\ldots (15.8)$$

Where **R** is constant of proportionally and is called General gas constant or universal gas constant and does not depend on the quantity of gas in the sample. If P is measured in Nm^{-2} V in m^3 and T in Kelvin then the volume of universal gas constant is $\mathbf{R = 8.314 J\ mol^{-1}, k^{-1}}$ Above equation is written as:

$$PV = nRT \quad \ldots\ldots\ldots\ldots\ldots\ldots (15.9)$$

Worked Example 15.2

A cylinder contains 0.80 dm^3 of nitrogen gas at a pressure of 1.2 atmosphere (1 atm = 1.01 10^5 N/m^2). A piston slowly starts compressing the gas to a pressure of 6.0 atm. The temperature of the gas remains constant. Calculate the final volume of the gas. (The temperature of the gas is constant and that its mass is fixed).

Solution:

Step 1: Write the known quantities and point out quantities to be found.

$P_1 = 1.2$ atm, $V_1 = 0.80 dm^3$, $P_2 = 6.0$ atm, $V_2 = ?$

We don't worry about the particular units of pressure and volume being used here, so long as they are the same on both sides of the equation. The final value of V_2 will be in dm^3, because V_1 is in dm^3

Step 2: Write down the formula and rearrange if necessary.

$P_1V_1 = P_2V_2$

Substitute the values in equation and calculate.

$1.2 \text{ atm} \times 0.8 dm^3 = 6.0 \text{ atm} \times V_2$

$$V_2 = \frac{1.2 \text{ atm} \times 0.8 dm^3}{6.0 \text{ atm}} = 0.16 \text{ dm}^3$$

Result: The volume of the gas is reduced to 0.16dm3or 1/5th of original volume so the pressure increases by a factor of 5.

Worked Example 15.3

Find the mass of the air within the car tyre, given that it holds $0.020m^3$ of air at 27°C and a pressure of 3.0×10^5 N/m². The molecular mass of air is 28.8g/mol.

Solution:

Step 1: Write the known quantities and quantities to be found.

$P = 3.0 \times 10^5 N/m^2$ R = 8.314J mol–1 k–1

T = 27°C = 300K i. No. of moles of air = n =?

$V = 0.020m^3$ ii. Mass of the air = m =?

Molar mass of air = 28.8g/mol.

Step 2: Write down the formula and rearrange if necessary.

$$n = \frac{PV}{RT} = \frac{3.0 \times 10^5 \times 0.020}{8.314 \times 300} = 2.41 mol$$

Step 3: Now we calculate the mass of air.

mass = number of moles × molar mass

m = 2.41 mol × 28.8g/mol = 69.4g ≅ 69g

Result: i. 2.41 mol ii. 69g

Self-Assessment Questions:

1. How does the pressure of a gas change if its volume is halved at constant temperature?
2. What happens to the volume of a gas when its temperature is increased while keeping the pressure constant?
3. What is the relationship between the volume and the number of moles of a gas at constant temperature and pressure?

15.3 Kinetic theory of gases:

We discussed the properties of an ideal gas by using such macroscopic variables such as pressure, volume and temperature. Such large-scale properties can be related to a description on a microscopic scale, where matter is treated as collection of molecules. Now we begin by developing a microscopic model of an ideal gas called Kinetic Theory of gases. The first step in the construction of a theory is to setup some sort of model which is simple enough to be treated analytically.

15.3.1 Basic Postulates of Kinetic Theory of Gases:

The basic postulates of Kinetic Theory of gases are as under

1. A gas contains a very large number of particles called molecules. Depending on the gases each molecule consists of an atom or a group of atoms.
2. A finite volume of a gas consists of very large number of molecules. This assumption is justified by experiments. At standard conditions there are 3×10^{25} molecules per cubic meter.
3. The size of the molecules is much smaller than separation between molecules; it is about 3×10^{-10}m.

4. The molecules move in all directions with various speeds collide elastically with one another and with the walls of the container.
5. The molecules exert no forces on one another except during collisions. In the absence of the external forces, they move freely in straight lines.
6. Laws of mechanics are assumed to be applicable to the motion of molecules.

15.3.2 Pressure of Gas:

The pressure exerted by a gas is merely the momentum transferred to the walls of the container per second per unit area due to the continuous collisions of molecules of the gas. In order to calculate the pressure of an ideal gas from Kinetic Theory. Let us consider a cube having side length L whose walls are perfectly elastic contains N number of molecules each of mass m as shown in figure 15.6.

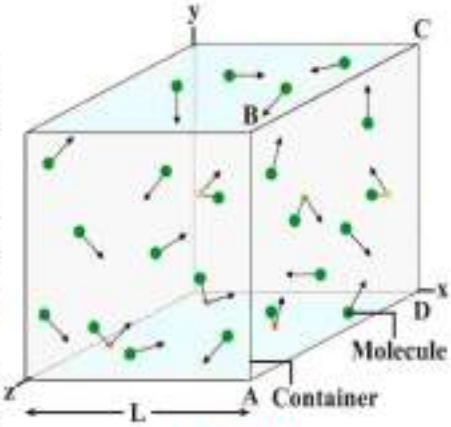


Figure 15.6
Shows gas molecules moving in random directions within container.

Consider a single molecule of mass m moving with velocity V_1parallel to x-axis.It moves back and forth, colliding at regular intervals with the ends of the box and thereby contributing to the pressure of the gas. A molecule which has a velocity V_1 can be resolved into three rectangular components V_1x, V_1y and V_1z parallel to three co-ordinates axis x, y and z.

A molecule which collides with the face ABCDA of the cube, it will rebound elastically in opposite direction, such that x-component of the velocity V_1x, is reversed, the $V_1y$and V_1z remain unaffected. Therefore, the momentum before collision is mV_1x and after collision is $-m V_1x$ causing a change of momentum.

Change in momentum= $P_i.P_f = -mV_1x - (-mV_1x) = mV_1x + mV_1x$

Change in momentum = $2mV_1x$ …….. 15.10

After recoil the molecule travels to opposite face and collides with it, rebounds and travels back to the face ABCDA after covering a distance 2L. The time Δt between two successive collisions with face ABCDA is:

$$\Delta t = \frac{2L}{V_1x} \quad \text{…….. (15.11)}$$

Now we can find the force that this one molecule exerts on face ABCDA, using Newton's 2nd law of motion. This says that the rate of change of momentum of the molecule is equal to force applied by the wall. According to Newton's 3rd law of motion, force F_1 exerted the molecule on face ABCDA is equal but opposite.

$$\text{Force} = F_1 = \frac{\text{change in momentum}}{\text{time taken}} = \frac{\Delta P}{\Delta t}$$

$$F_1 = \frac{2(mV_1x)}{\frac{2L}{V_1x}} = \frac{mV_1^2x}{L}$$

Similarly, the forces due to all other molecules can be determined. Thus, the total x-directed F due to N number of molecules of the gas moving with velocities V_1, V_2, V_3,……,V_n is:

$$F = F_1 + F_2 + F_3 + \ldots\ldots\ldots\ldots\ldots + F_n$$

$$F = \frac{mV_1^2x}{L} + \frac{mV_2^2x}{L} + \frac{mV_3^2x}{L} + \ldots\ldots\ldots\ldots\ldots + \frac{mV_n^2x}{L}$$

As pressure is normal force per unit area, hence pressure P on the face perpendicular to x-axis is:

$$P = \frac{F}{A} = \frac{F}{L^2}$$

$$P = \left(\frac{mV_1^2x}{L} + \frac{mV_2^2x}{L} + \frac{mV_3^2x}{L} + \ldots\ldots\ldots + \frac{mV_n^2x}{L}\right)/L^2$$

$$P = \frac{m}{L^3}(V_1^2x + V_2^2x + V_3^2x + \ldots\ldots\ldots + V_n^2x)\ldots\ldots\ldots\ldots(15.12)$$

The number of molecules in unit volume n_v is $\frac{N}{L^3}$. Where N is the total number of molecules. Therefore:

$L^3 = \frac{N}{n_v}$ and substituting this value in eq: (15.12)

$$P = mn_v\left(\frac{V_1^2x+V_2^2x+V_3^2x+\ldots\ldots\ldots+V_n^2x}{N}\right)\ldots\ldots\ldots\ (15.13)$$

Where mn_v is the mass per unit volume which we call density ρ and $\frac{V_1^2x+V_2^2x+V_3^2x+\ldots\ldots\ldots+V_n^2x}{N}$ is the average value of $\overline{V_x^2}$. for all the molecules in the container, we call this average square velocity $\overline{V_x^2}$. The square root of $\overline{V_x^2}$. is referred as V_{rms}. Eq: 15.13 can be written as:

$$P = \rho\overline{V_x^2}\ .\ \ldots\ldots\ldots\ldots\ldots\ldots\ (15.14)$$

The terms $\overline{V_x^2}$. is only one component of the total velocity. Since $\overline{V^2} = \overline{V_x^2} + \overline{V_y^2} + \overline{V_z^2}$ on the average. $\overline{V_x^2} = \overline{V_y^2} = \overline{V_z^2}$ due to randomness of the molecular motion.

$\overline{V^2} = 3\overline{V_x^2}$. and $\overline{V_x^2} = \frac{1}{3}\overline{V^2}$

Substituting this value into the above equation, we find that:

$$P = \frac{1}{3}\rho\overline{V^2}\ \ldots\ldots\ldots\ldots\ldots\ldots\ (15.15)$$

15.3.3 The relation between Kinetic Energy of Molecules and Temperature:

The relation between molecular Kinetic Energy and temperature can be derived by using eq: 15.15

$$P = \frac{1}{3}\rho\overline{V^2}$$

$$P = \frac{1}{3}mn_v\overline{V^2} \text{ since } \ell = (mn_v)\ldots\ldots\ldots\ldots\ (15.16)$$

Since n_v represents the number of molecules per unit volume $n_v = \frac{N}{V}$ Eq: (15.16) can be written as

$$PV = \frac{1}{3}m\frac{N}{V}\overline{V^2}\ldots\ldots\ldots\ldots\ (15.17)$$

Now we can compare the equation (15.17) with this ideal gas equation PV = nRT. The left-hand sides are the same. So, the two right-hand sides must also be equal.

$\frac{1}{3}Nm\overline{V^2} = nRT$ ..... (15.18)

Substituting $n = \frac{N}{N_A}$ and multiplying both sides by $\frac{3}{2}$, we obtain the relation.

$\frac{3}{2} \times \frac{1}{3}Nm\overline{V^2} = \frac{3}{2} \times \frac{N}{N_A}RT$

$\frac{1}{2}m\overline{V^2} = \left(\frac{3}{2}\right)\left(\frac{R}{N_A}\right)T$

Since $\frac{R}{N_A} = k$ (Boltzmann constant)

$k = 1.38 \times 10^{-23}$ J/Molecule k

Hence $\frac{1}{2}m\overline{V^2} = \frac{3}{2}kT$ ..... (15.19)

$K.E_{avg} = \frac{3}{2}kT$

DO YOU KNOW?

K (Boltzmann constant) is universal gas constant per molecule of a gas R is a universal gas. Constant per mole of gas.

The mean translational Kinetic energy of a molecule of an ideal gas is proportional to the absolute temperature.

Hence, mean translational Kinetic energy of a molecule $\propto$ T.

Worked Example 15.4

The atoms in a gas have mean translational (K.E) equal to 5.0×10^{-21}J. Calculate the temperature of the gas in K and °C.

Solution:

Step 1: Write the known quantities and point out quantities to be found

Mean K.E = 5.0×10^{-21}J

K = 1.38×10^{-23}J/molecule. K

(i). T_k = ?

(ii). T_c = ?

Step 2: Write down the formula and rearrange if necessary.

Mean K.E = $\frac{3}{2}kT$

$T = \frac{\text{mean K.E}}{k} \times \frac{2}{3} = \frac{5.0 \times 10^{-21}}{1.38 \times 10^{-23}} \times \frac{2}{3}$

T = 242K

Thus, temperature in centigrade scale is:

T = K – 273

T = 242 – 273 = – 31°C.

Result: (i). 242K, **(ii).** –31°C.

Self-Assessment Questions:

1. Describe the motion of molecules in a gas according to the kinetic molecular theory.
2. What is the relationship between temperature and molecular motion in a gas?

SUMMARY

- Temperature is a measure of the average kinetic energy of particles in a substance.
- The triple point of water is the temperature (0.01°C) and pressure (611.73 Pascal's) at which water coexists in all three phases (solid, liquid, and gas).
- The general gas law states that $PV = nRT$, relating pressure, volume, number of moles, and temperature of a gas.
- Boyle's Law: $P_1V_1 = P_2V_2$ (pressure and volume are inversely proportional at constant temperature)
- $V_1/T_1 = V_2/T_2$ (volume and temperature are directly proportional at constant pressure)
- $V = RT/P$ (volume and number of moles are directly proportional at constant temperature and pressure)
- KTG (Kinetic Theory of Gases) Postulates:
  1. Gases consist of tiny particles (molecules) that are in constant motion.
  2. The molecules are very small compared to the distance between them.
  3. The molecules are in constant random motion.
  4. The molecules collide with each other and the container walls.
- Molecular movement and pressure: The pressure exerted by a gas is caused by the collisions of molecules with the container walls, and is related to the temperature and volume of the gas by the equation $P = (2/3)\ nkT$.
- The kinetic energy of molecules is directly proportional to the temperature of the gas, and is described by the equation $KE = (3/2)\ kT$.

Unit 16 First Law of Thermodynamics

Teaching Periods 08 **Weightage %** 05

The Thar Coal Power Plant, located in the Thar Desert of Sindh, Pakistan, is a significant project aimed at utilizing the large reserves of lignite coal found in the region. The Thar Coal Power Plant is an example of how the first law of thermodynamics is applied in real-world energy systems.

In this unit student should be able to:

- Describe that heat flow and work are two forms of energy transfer between systems and calculate heat being transferred.
- Relate rise in temperature of a system increases its internal energy.
- Explain that internal energy is determined by the state of the system and that it can be expressed as the sum of the random distribution of kinetic and potential energies associated with the molecules of the system.
- Define thermodynamics and various terms associated with it.
- Calculate work done by a thermodynamic system during a volume change.
- Describe the first law of thermodynamics expressed in terms of the change in internal energy, the heating of the system and work done on the system.
- Explain that first law of thermodynamics expresses the conservation of energy.
- Describe the applications of first law of thermodynamics with the help of equations and graphs
- Solve the problems using the equations of first law of thermodynamics
- Define the terms specific heat and molar specific heats of a gas.
- Apply first law of thermodynamics to derive $C_p - C_v = R$.
- Solve the problems using equations of specific heat

Introduction:

In this unit, we explore key concepts in thermodynamics: heat, work, internal energy, the First Law of Thermodynamics, and molar specific heat. These concepts form the basis for understanding energy transfer and conservation within systems. Heat and work are fundamental modes of energy transfer, influencing changes in a system's internal energy. This unit provides a comprehensive overview, essential for grasping energy transformations and their applications in diverse fields.

16.1.1 Heat and Work:

Work and thermal energy are linked by the broader framework of energy conservation. The study of heat has evolved significantly over centuries. As shown in table 16.1.

Table 16.1: History of heat energy

18th Century	Isaac Newton	➢ Proposed the concept of "caloric" (a hypothetical fluid thought to be the substance of heat). ➢ Influenced future research.
18th Century	Joseph Black	➢ Introduced the concept of latent heat. ➢ Distinguished between heat and temperature.
18th Century	Count Rumford (Benjamin Thompson)	➢ Challenged the caloric theory of heat ➢ Demonstrated that heat is a form of motion, not a material substance
19th Century	Sadi Carnot	➢ Laid the foundations of thermodynamics ➢ Described the Carnot cycle
19th Century	James Prescott Joul	➢ Established the mechanical equivalent of heat ➢ Showed that energy can be converted from one form to another
19th Century	Rudolf Clausius and William Thomson (Lord Kelvin)	➢ Developed the concept of entropy ➢ Formulated the second law of thermodynamics
20th Century	Albert Einstein	➢ Provided evidence for the atomic theory of matter ➢ Helped establish the kinetic theory of heat
Modern Developments:		➢ Advances in statistical mechanics and quantum mechanics ➢ Connected heat to the behavior of atoms and molecules

These discoveries laid the foundation for modern thermodynamics, which explains heat as the kinetic energy of particles in motion.

Work can lead to changes in the internal energy of a system, contributing to thermal energy. The first law of thermodynamics encapsulates this relationship, emphasizing the interplay between heat transfer, work and changes in internal energy within a system.

Derivation:

The volume of gas is at constant pressure. This means the force F exerted by the gas on the piston is equal to:

$$F = p \times A$$

Where: p = pressure of the gas (Pa) and A = cross-sectional area of the cylinder (m^2)

The definition of work done is:

$$W = F \times S$$

Where: F = force (N) and S = displacement in the direction of force (m)

The displacement of the gas d multiplied by the cross-sectional area A is the increase in volume ΔV of the gas:

$$W = P \times A \times S \ldots\ldots 16.4$$

This gives the equation for the work done when the volume of a gas changes at constant pressure:

$$W = P\Delta V$$

Where: ΔV = increase in the volume of the gas in the piston when expanding (m^3)

This is assuming that the surrounding pressure p does not change as the gas expands, This will be true if the gas is expanding against the pressure of the atmosphere, which changes very slowly, When the gas expands (V increases), work is done by the gas. When the gas is compressed (V decreases), work is done on the gas.

Worked Example 16.1

The volume occupied by 1.00 mol of a liquid at 50 °C is 2.4×10^{-5} m^3. When the liquid is vaporized at an atmospheric pressure of 1.03×10^5Pa, the vapor has a volume of 5.9×10^{-2} m^3.The latent heat to vaporize 1.00 mol of this liquid at 50 °C at atmospheric pressure is 3.48×10^4 J. Determine change in internal energy ΔU of the system.

Solution:

Step 1: Write down the first law of thermodynamics

$\Delta U = Q - W$

Step 2: Write the value of heating q of the system

This is the latent heat, the heat required to vaporize the liquid = $\mathbf{3.48 \times 10^4}$ **J**

Step 3: Calculate the work done W

$$W = P.\Delta V$$

$\Delta V = V_f - V_i = 5.9 \times 10^{-2} - 2.4 \times 10^{-5} = 0.058976\ m^3$

p = atmospheric pressure = 1.03×10^5 Pa

$W = (1.03 \times 10^5) \times 0.058976 = 6074.528 = 6.07 \times 10^3$ J

Since the gas is expanding, this work done is negative

$W = -6.07 \times 10^3$ J

Step 4: Substitute the values into first law of thermodynamics

$\Delta U = 3.48 \times 10^4 \quad (-6.07 \times 10^3) = 28\ 730 = 29\ 000$ J

$T_f = = \frac{9}{5} \times 36.88 + 32$

$T_f = 98.4°F$

Result: Result: $T_f = 98.4°$

Worked Example 16.2

When a balloon is inflated, its rubber walls push against the air around it. Calculate the work done when the balloon is blown up from 0.015 m^3 to 0.030 m^3. Atmospheric pressure = 1.0×10^5 Pa.

Solution:

Step 1: write the known quantities and point out quantities to be found

$V_1 = 0.015\ m^3$

$V_2 = 0.030\ m^3$

$P = 1.0 \times 10^5$ Pa.

W= ?

Step 2: Write down the equation for the work done by a gas

$$W = P\Delta V$$

Step 3: Now First, calculate ΔV:

$$\Delta V = V_f - V_i$$

$\Delta V = 0.030\ m^3 - 0.015\ m^3 = 0.015\ m^3$

Step 4: Now, substitute the values into the formula:

$$W = -P\Delta V$$

$W = (1.0 \times 105) \times 0.015 = 1500$ J

Result: The work done is **1500 J**

16.3.3: The first law of thermodynamics, internal energy and the heating of the system& conservation of energy:

First law of thermodynamics:

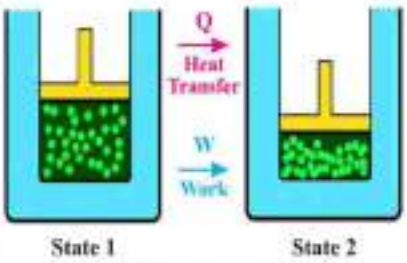
The first law of thermodynamics is based on the idea that energy can neither be created nor destroyed in any thermodynamic system. So, it is a particular form of the law of conservation of energy, which deals only with the heat energy.

"This law states that every thermodynamic system possesses a state variable (U) called the internal energy".

In any thermodynamic process, when heat energy (ΔQ) is added to a system, this energy appears as an increase in the internal energy (ΔU) stored in the system plus the work done (ΔW) by the system on its surroundings.

$$\Delta Q = \Delta U + \Delta W \quad \text{.....} \quad 16.5$$

ΔQ is taken positive when heat enters the system and negative when it leaves the system. ΔU is taken positive when the temperature of the system rises, while it is negative when the temperature of the system decreases.



$\Delta U = \Delta Q - W$

Figure 16.6 states of system

Adiabatic Compression:

If the gas is compressed by suddenly increasing the pressure on the piston, the gas will be heated, so that temperature will increase. Therefore, the internal energy of the gas will increase.

According to the first law of thermodynamics;

$$\Delta Q = \Delta U + \Delta W$$
$$0 = \Delta U + \Delta W$$
$$\Delta U = -\Delta W$$

Thus, an increase in the internal energy of the system in an adiabatic process is equal to the work done on the system.

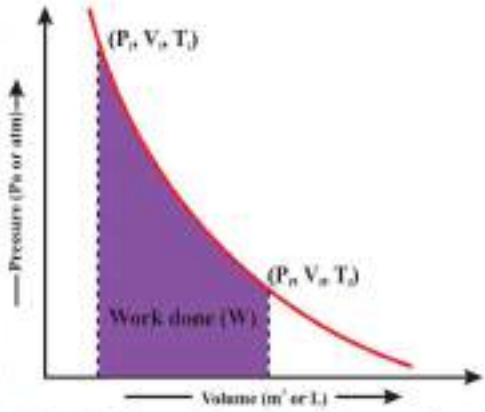


Figure 16.11 graph for adiabatic Compression

Above equation can be written as;

$$\Delta W = -\Delta U \ldots\ldots\ldots 16.11$$

This means that if the system does the work, then in adiabatic process, the work is done at the cost of internal energy.

Self-Assessment Questions:

1. What is the first law of thermodynamics and how is it expressed mathematically?
2. How does the first law of thermodynamics relate to the conservation of energy?

Worked Example 16.3

The normal human body temperature is 36.88°C. What is this temperature on Fahrenheit scale? A gas confined in a cylinder undergoes an expansion from an initial volume of 2.0 L to a final volume of 4.0 L against a constant external pressure of 2.0 atm. Calculate the work done by the gas during this process.

Solution:

Step 1: Write the known quantities and point out quantities to be found.

Initial volume $V_i = 2.0$ L

Final volume $V_f = 4.0$ L

Constant external pressure $(P) = 2.0$ atm

Step 2: Write the formula and rearrange if necessary.

The work done by a gas against a constant external pressure can be calculated using the formula:

$$W = -P\Delta V$$

where: W is the work done,

P is the constant external pressure, and

ΔV is the change in volume.

Step 3: First, calculate ΔV:

$$\Delta V = V_f - V_i$$
$$\Delta V = 4.0L - 2.0L = 2.0L$$

Step 4: Now, substitute the values into the formula:

$$W = -P\Delta V$$
$$W = -(2.0atm) \times (2.0L)$$
$$W = -4.0atm \cdot L$$

Result: The negative sign indicates that work is done on the gas (compression) rather than by the gas (expansion). The final result is −4.0atm·L of work done on the gas during this expansion process.

Remember that the unit of work in the International System of Units (SI) is the joule (J), so you may want to convert the result to joules if needed (1 atm·L = 101.325 J).

Self-Assessment Questions:

1. How does an adiabatic process differ from an isothermal process in thermodynamics?
2. Why is an isothermal process important in thermodynamics, and how does it affect the internal energy of a system?
3. Why is an isobaric process significant in thermodynamics, and how does it affect the volume of a gas?

16.4.1 Heat capacity:

When heat is absorbed by a body, its temperature increases, and when heat is lost, its temperature decreases. Temperature represents the total kinetic energy of the particles comprising the body. Therefore, when heat is absorbed, it increases the kinetic energy of these particles, thereby raising the temperature. Consequently, the change in temperature ΔT is directly proportional to the amount of heat transferred q.

The formula $Q = nC\Delta T$ expresses the relationship where q is the heat transferred, n is the amount of substance (often expressed in moles), C is the molar heat capacity, and ΔT is the change in temperature. Molar heat capacity C refers to the amount of heat energy required to change the temperature of 1 mole of a substance by 1 unit. It is influenced by the substance's properties such as its nature, size, and composition.

Specific Heat:

Specific heat, often denoted by C orc, is a measure of the heat energy required to raise the temperature of a unit mass (usually one gram or one kilogram) of a substance by one degree Celsius (or one Kelvin). The formula for specific heat (c) is given by:

$$\boldsymbol{Q = mc\Delta T} \ldots\ldots\ldots \mathbf{16.12}$$

where:

Q is the heat energy absorbed or released,
m is the mass of the substance,
c is the specific heat, and
ΔT is the change in temperature.

Specific heat is an intensive property, meaning it does not depend on the quantity of the substance.

Molar Specific Heat:

Molar specific heat, denoted as C_m, measures how much heat energy is needed to raise the temperature of one mole of a substance by one degree Celsius or Kelvin. It's calculated using the formula:

$$Q = nC_m \Delta T \ldots\ldots 16.13$$

where:

Q is the heat transferred,
n is the number of moles,
C_m is the molar specific heat,
ΔT is the temperature change.

Molar specific heat is an intensive property and relates to specific heat C through the equation $C_m = nC$

In thermodynamics, both specific heat and molar specific heat are vital for quantifying how heat energy changes with temperature in substances. These values vary depending on the substance, reflecting each substance's unique ability to absorb or release heat.

There are two types of molar heat capacity:

C_P (molar heat capacity at constant pressure)
C_V (molar heat capacity at constant volume)

The relationship between CP and CV for an Ideal Gas:

From the equation $q = n\,C\,\Delta T$, we can derive:

At constant pressure $q_p = n\,C_p \Delta T$ This value corresponds to the change in enthalpy, expressed as: $q_P = n\,C_P \Delta T = \Delta H$

Similarly, at constant volume $Q_v = n\,C_v \Delta T$

This value corresponds to the change in internal energy, expressed as:

$$Q_v = n\,C_v\,\Delta T = \Delta U$$

For one mole of an ideal gas (n=1)

$$\Delta H = \Delta U + \Delta(PV) = \Delta U + \Delta(RT) = \Delta U + R\,\Delta T \ldots\ldots 16.14$$

Therefore, we have: $\Delta H = \Delta U + R\,\Delta T$

Substituting the expressions for

$$C_p \Delta T = C_v\,\Delta T + R\,\Delta T$$

$$C_p = C_p - C_v = R \ldots\ldots 16.15$$

This equation shows the relationship between the molar heat capacities at constant pressure C_P and constant volume C_V for an ideal gas, where R is the universal gas constant.

Worked Example 16.4

Calculate the molar specific heat capacity of a gas when 2 moles of the gas absorb 1500 J of heat energy, and its temperature increases by 25 degrees Celsius.

Solution:

Step1: write the known quantities and point out quantities to be found

$Q = 1500J$, $n = 2mol$, and $\Delta T = 25°C$.

Step 2: Write the formula and rearrange if necessary

The heat transfer (Q) is given by:

$Q = n \cdot Cm \cdot \Delta T$

Let's rearrange the formula to solve for Cm:

$Cm = n \cdot \Delta TQ$

Step 3: Now, substitute the values into the formula:

$Cm = 1500J/2mol \cdot 25°C$

$Cm = 30J/mol \cdot °C$

Therefore, the molar specific heat capacity of the gas is 30J/mol·°C.

Result: 30J/mol·°C.

Worked Example 16.5

A sample of water with a mass of 200 grams is heated, and its temperature rises from 25°C to 45°C. Calculate the heat transferred to the water. Given the specific heat capacity of water (c) is 4.18 J/g°C.

Solution: Write the known quantities and point out quantities to be found

$m = 200g$

$c = 4.18J/g°C$

Initial temperature (T1) = 25°C

Final temperature (T2) = 45°C

Calculate the temperature change (ΔT): $\Delta T = T2 - T1 = 45°C - 25°C = 20°C$

Step 2: Write the formula and rearrange if necessary

heat transfer equation $Q = mc\Delta T$ to calculate the heat transferred,

Step 3: Now, substitute the values into the formula:

Substitute the values: Q=(200g)×(4.18J/g°C)×(20°C)

Calculate the result: Q=8360J

Result: The heat transferred to the water is 8360 joules.

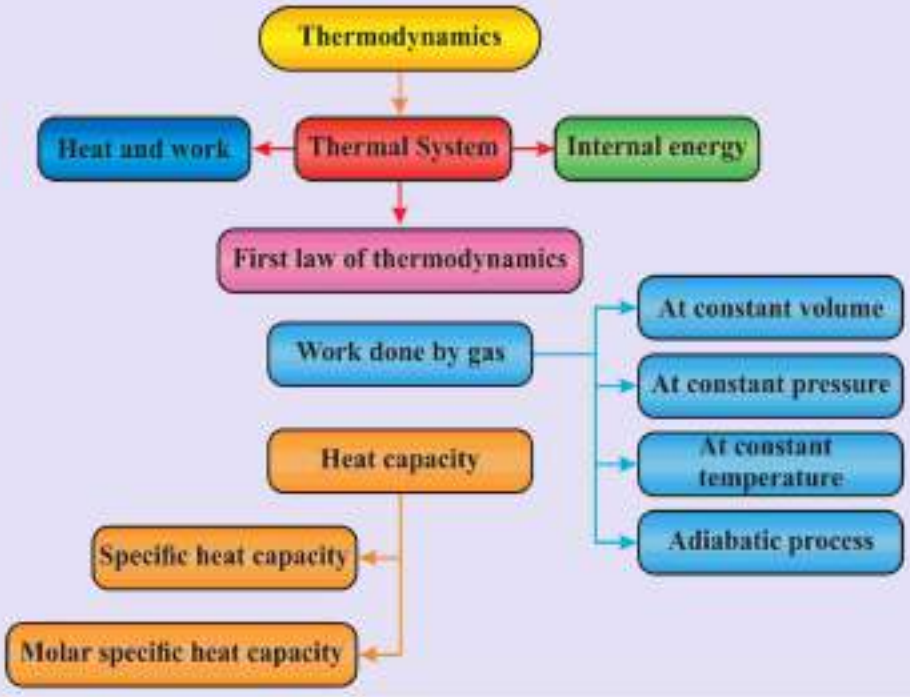
Self-Assessment Questions:

1. How does heat capacity relate to the amount of energy required to change the temperature of a substance?
2. Explain why different materials have different heat capacities.
3. How does the heat capacity of a substance change with temperature?

SUMMARY

- Heat (Q): Transfer of thermal energy between a system and its surroundings.
- Work (W): Mechanical energy transferred to or from a system through forces acting on it.
- Sign conventions: Heat added to the system and work done by the system are positive, while heat removed from the system and work done on the system are negative.
- Equation: $\Delta U = Q - W$, where ΔU is change in internal energy, Q is heat added to the system, and W is work done by the system.
- Internal energy (U): Total energy of a system, including kinetic and potential energies.
- First Law of Thermodynamics: Conservation of energy principle in thermodynamic systems.
- Energy cannot be created or destroyed, only transferred or converted from one form to another.
- Applications: Understanding energy transformations in various processes such as heating, cooling, phase changes, and chemical reactions.

CONCEPT MAP



3. Distinguish between the work done by a system and the heat exchanged with the surroundings in the context of the first law.
4. Give an example daily life that illustrates the principles of the first law of thermodynamics.
5. Explain the role of the system and its surroundings in the context of the first law of thermodynamics.
6. How does heat capacity relate to the amount of energy required to change the temperature of a substance?

Section (C): ERQs (Long Answered Questions):

1. Provide the mathematical expression of the first law of thermodynamics and explain each term.
2. Describe what happens to the internal energy of a system in an adiabatic process, and why.
3. What is meant by the internal energy of a system, and how does it change during various thermodynamic processes?
4. Explore the concept of positive and negative work done by a gas, depending on whether the gas is expanding or compressing.
5. Explain the concept of heat capacity and its significance in thermodynamics. Discuss how heat capacity relates to the ability of a substance to store thermal energy and its implications for temperature changes.

Section (D): Numerical:

1. A gas undergoes isothermal expansion at a constant temperature of 300 K. If the gas absorbs 500 J of heat during the process, calculate the work done by the gas. **(500 J)**
2. A piston compresses a gas adiabatically. If the initial volume is 0.02 m^3 and the final volume is 0.01 m^3, and the initial pressure is 200 kPa, determine the final pressure. Assume the gas behaves ideally. **(400 kPa)**
3. A system undergoes an isobaric process where the pressure is kept constant at 150 kPa. If the volume increases from 0.05 m^3 to 0.08 m^3,calculate the heat added to the system. **(600 J)**
4. During an isochoric process, the internal energy of a gas increases by 300 J. If no work is done, determine the heat added to the system. **(300 J)**
5. A gas undergoes a cyclic process, starting at point A with a volume of 0.02 m^3, going to B (isochoric heating), then to C (isothermal expansion), and finally back to A. If the heat added during isothermal expansion is 1000 J and the heat rejected during isochoric heating is 500 J, calculate the net work done by the system. **(500 J)**
6. A gas expands from 0.03 m^3 to 0.06 m^3 against a constant pressure of 100 kPa. Calculate the work done in both a reversible and an irreversible process, and compare the results. **(300 J and 600 J)**
7. A 50 g piece of copper at 100°C is placed in 200 g of water at 20°C. If the final temperature of the system is 30°C, calculate the specific heat capacity of copper. (Specific heat capacity of water = 4.18 J/g°C) **(0.39 J/g°C)**
8. How much heat is required to raise the temperature of 1 kg of lead from 25°C to 100°C? (Specific heat capacity of lead = 0.128 J/g°C). **(0240J).**

Unit 17 Second Law of Thermodynamics

Teaching Periods 10 **Weightage %** 06

The neem tree, like all living systems, increases entropy in its surroundings through various processes, such as:

1. Transpiration: Releasing water vapors into the air, increasing humidity and entropy.

2. Photosynthesis: Absorbing sunlight and converting it into chemical energy, releasing excess energy as heat and increasing entropy.

These processes illustrate the second law of thermodynamics, where living systems increase entropy in their surroundings while maintaining their own organization and complexity.

In this unit student should be able to:

- State and explain second law of thermodynamics.
- State the working principle of heat engine.
- Describe the concept of reversible and irreversible processes.
- Describe the working of petrol engine and diesel engine.
- Explain the working principle of Carnot's engine
- Explain that the efficiency of a Carnot engine is independent of the nature of the working substance and depends on the temperatures of hot and cold reservoirs. Solve problems to find out the efficiency of heat engine
- Describe that refrigerator is a heat engine operating in reverse as that of an ideal heat engine and find its efficiency
- Solve problems to find out the efficiency of a refrigerator
- Describe that change in entropy is positive when heat is added and negative when heat is removed from the system.
- Explain that increase in entropy means degradation of energy.
- Show that energy is degraded during all natural processes.
- Identify that system tend to become less orderly over time
- Solve problems using the equation of entropy

Introduction:

The first law of thermodynamics is a particular form of the law of conservation of energy. However, there are some limitations regarding the transformation of heat energy into mechanical energy, which are covered by the 2nd law of thermodynamics.

17.1.1 Second law of thermodynamics:

Lord Kelvin and Rudolf Clausius formulated the second law of thermodynamics. It consists of two statements.

Kelvin Statement:

According to this statement ***"It is impossible to construct an engine, operating continuously in a cycle that can take heat from a source and coverts completely into work".***

According to Kelvin statement, an engine when converting heat into mechanical work cannot convert all of it into work. A part of heat must be rejected to a cooler reservoir, the exhaust as shown in figure 17.1(a).

In other words, it is impossible to device an engine that would have an efficiency of 100%, even though the first law of thermodynamics will be satisfied.

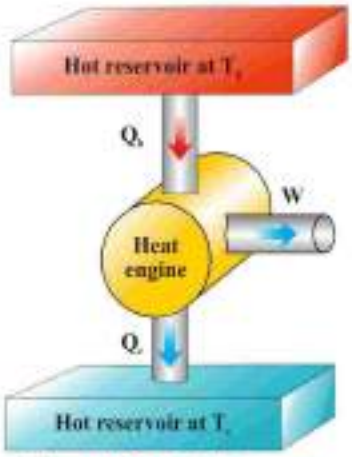


Figure 17.1 (a) the heat engine

Clausius Statement:

It is fact that no one has ever built a refrigerator which will work without a supply of energy. A refrigerator is essentially a machine for conveying heat from one body at a low temperature to another at a high temperature as shown in figure: 17.1(b). According to Clausius statement

"It is impossible to cause heat to flow from a cold body to a hot body without the expenditure of energy".

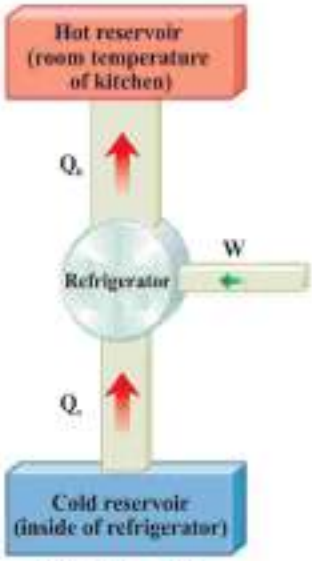


Figure 17.1 (b) the refrigerator

17.2.1 Working principle of heat engines:

Any device that transforms heat into mechanical energy (work) is called heat engine. The essentials of a heat engine are the furnace or hot body, the working substance and a condenser or cold body.

The simplest kind of engine to discuss is one in which the working substance undergoes a cyclic process, that is, a sequence of processes that eventually leaves the substance in the same state in which it started. In a steam turbine the water is recycled and used over and over. Internal combustion

engines do not use the same air again but we can still analyze them in terms of cyclic processes that approximate their actual operation.

All the heat engines absorbs heat from a source at a relatively low temperature performs some mechanical work and discards some heat at a low temperature the engine treats extra heat as waste. It gets rid of this waste heat through the exhaust pipes and the cooling system in internal combustion engines.

DO YOU KNOW?

Each Q is positive when heat is transferred from a reservoir into the working substance and Q is negative when the reverse happens

In a heat engine, Q_1 is positive but Q_2 is negative, representing heat leaving the working substance.

We can represent the energy transformation in a heat engine by the flow diagram fig: 17.2. The engine itself is represented by the circle. The amount of heat suppliedQ_1 to the engine by the hot reservoir is proportional to the cross section of the incoming pipeline at the top of the diagram. The cross section of the outgoing of the pipeline at the bottom is proportional to the magnitude $|Q_2|$of the heat discarded in the exhaust.

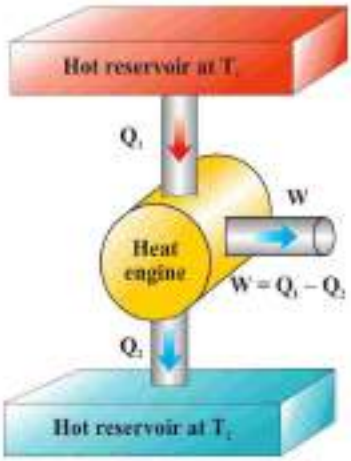


Fig: 17.2 Heat Engine

The branch line to the right represents the portion of the heat supplied that the engine converts to mechanical work W.

When an engine repeats the same cycle over and over, Q_1 and Q_2 represent the quantities of heat absorbed and rejected by the engine during one cycle. The net heat absorbed per cycle is

$$\Delta Q = Q_1 - Q_2$$

Let Q_1 be the heat absorbed by an engine from a high temperature reservoir and Q_2 be the heat rejected by the engine to low temperature reservoir or sink. The rest of heat energy is converted into useful work i.e.

$$W = Q_1 - Q_2 \ldots\ldots 17.1$$

Efficiency of heat engine:

"The thermal efficiency "η" of a cycle heat engine is defined to the ratio of the network 'W' done by the engine in each cycle to the heat absorbed Q_1 in each cycle".

$$Efficiency\ (\eta) = \frac{output}{input} = \frac{W}{Q_1} = \frac{Q_1 - Q_2}{Q_1} = 1 - \frac{Q_2}{Q_1} \ldots\ldots 17.2$$

For 100% efficiency $Q_2 = 0$ i.e. no heat is rejected to low temperature reservoir (sink) and whole heat absorbed is converted into work. But experimental facts show that no cyclic engine can achieve an efficiency of 100%.

17.2.2 Reversible and Irreversible Processes:

To understand the nature of theoretical heat engine, we must first examine the concept of reversible and irreversible process. In a reversible process, the system undergoing the process

can be returned to its initial condition along the same path on a PV diagram and every point along this path is an equilibrium state. A process that does not satisfy these requirements is irreversible. All natural processes are known to be irreversible.

Consider an adiabatic process where a gas in a thermally insulated container as shown in fig: 17.3. A membrane separates the gas from a vacuum. When the membrane is burst, the gas expands freely into the vacuum. The system occupies a greater volume after the expansion. No heat energy is transferred to or from the container.

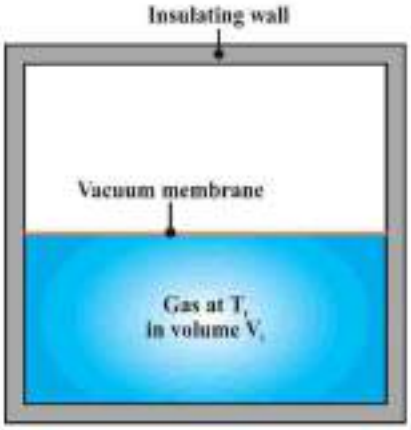


Figure 17.3 Thermally Insulated container

In this adiabatic process, the system undergoes changes without any heat exchange with its surroundings. Here is how the system changes. To make this process reversible, we must return the gas to its original volume and temperature without affecting the surroundings. If we try to reverse the process by compressing the gas back to its original volume using an engine, the surroundings change because external work is done on the system. The temperature increases during compression, which can be lowered by contact with an external energy reservoir. This returns the gas to its original state but affects the surroundings by adding energy. If this energy could drive the compression engine, the net energy transfer to the surroundings would be zero, making the process reversible. However, the Kelvin statement of the second law of thermodynamics specifies that the heat energy removed from the gas cannot be completely converted to mechanical work. Thus, the process is irreversible.

"A reversible process is one which can be retraced in exactly reverse order, without producing any change in the surroundings".

It is in equilibrium at all stages of operation. A succession of events which bring the system back to its initial condition is called a cycle. A reversible cycle is the one in which all the changes are reversible.

"The process which cannot be retraced in the backward direction by reversing the controlling factors is known as irreversible process"

It is in equilibrium only at initial and final stage. All changes which occur suddenly or which involved friction or dissipation of energy through conduction, convection or radiations are irreversible. An example of highly irreversible process is an explosion.

17.2.3 Petrol Engine:

Petrol engine is an internal combustion engine designed to run on volatile fuel such as petrol (gasoline), which has spark-ignition. In these engines air and fuel are generally mixed post-compression.

It works on the Otto cycle and the name comes from the German Engineer Nikolaus Otto (1876), who made the first working prototype. Although different engines may differ in

their construction technology but they are based on the principle of Carnot cycle. A typical four stroke petrol engine Fig. (17.4) also undergoes four successive processes in each cycle.

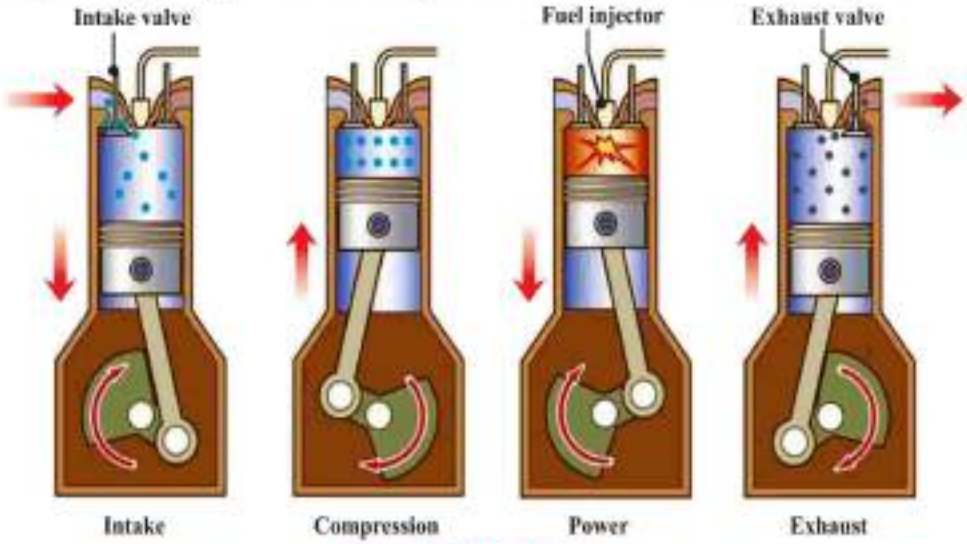


Figure17.4 Petrol Engine

1. The cycle starts on the intake stroke in which piston moves outward and petrol air mixture is drawn through an inlet valve into the cylinder from the carburetor at atmospheric pressure.
2. On the compression stroke, the inlet valve is closed and the mixture is compressed adiabatically by inward movement of the piston.
3. On the power stroke, a spark fires the mixture causing a rapid increase in pressure and temperature. The burning mixture expands adiabatically and forces the piston to move outward. This is the stroke which delivers power to crankshaft to drive the flywheels.
4. On the exhaust stroke, the outlet valves open. The residual gases are expelled and piston moves inward. The cycle then begins again. Most motorbikes have one cylinder engine but cars usually have four cylinders on the same crankshaft. The cylinders are timed to fire turn by turn in succession for a smooth running of the car. The actual efficiency of properly tuned engine is usually not more than 25% to 30% because of friction and other heat losses.

DO YOU KNOW?

No real engine operating between two energy reservoirs can be more efficient than a Carnot engine operating between the same two reservoirs in accordance with Carnot theorem.

17.2.4 Diesel Engine:

The diesel engine named after Rudolf Diesel (German inventor and Mechanical Engineer), is an internal combustion engine. He patented his original design in 1892. When the fuel comes into contact with high temperature, it ignites, creating energy that drives the piston

down transferring energy to the crankshaft. There are two classes of diesel engine: two strokes and four strokes. Most diesel engines generally use the four-stroke cycle.

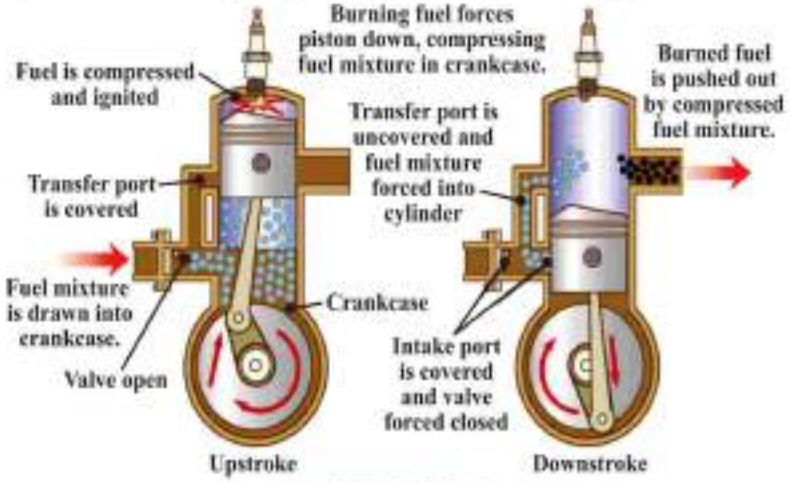


Fig: 17.5 Diesel Engine

No spark plug is needed in the diesel engine (fig: 17.5). Diesel is sprayed into the cylinder at maximum compression. Because air is at very high temperature immediately after compression, the fuel mixture ignites on contact with air in the cylinder and pushes the piston outward. The efficiency of diesel engine is about 35% to 40%.

17.3.1 The Carnot Engine:

A French Engineer named Sadi Carnot in 1824 described an ideal engine, now called a Carnot Engine, which is free from all sort of heat losses and friction. It consists of a hot body of infinite thermal capacity, a similar cold body, a perfect heat insulator and cylinder fitted with a piston enclosing any working substance. The cylinder has a heat conducting base and non-conducting walls and piston. The working substance is taken through the following cycle of operations known as the Carnot cycle. All real engines are less efficient than the Carnot engine because they do not operate through a reversible cycle. A cycle of a heat engine is completed when the properties of a system have returned to the original state. The operating cycle of the Carnot engine is called Carnot cycle. It consists of four processes, two isothermal and two adiabatic.

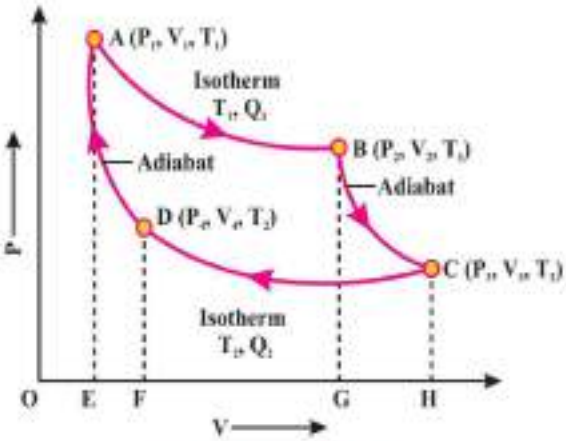


Figure 16.6 Carnot Cycle

1. In process A→B (fig.17.7 a) is isothermal expansion at temperature T_1. The cylinder is placed in thermal contact with an energy reservoir at temperature T_1.During the expansion, we decrease load on piston, the gas absorbs energy Q_1 from the reservoir through the base of the cylinder, volume of gas increases from V_1 to V_2 and does work W_{AB} in raising the piston.
2. In Process B→C (fig: 17.7 b), the base of the cylinder is replaced by a thermally non conducting wall and the gas expands adiabatically due to decrease in load on the piston; that is, no heat energy enters or leaves the system during the expansion, the temperature of the gas decreases from T_1 to T_2, volume increases from V_2 to V_3 and gas does work W_{BC} in raising the piston.
3. In process C→D (fig.17.7 c), the cylinder is placed in thermal contact with an energy reservoir at temperature T_2 is compressed isothermally at temperature T_2 by increasing the load on the piston. The volume decreases from V_3 to V_4. During this time, the gas expels energy Q_2 to the reservoir and the work done by the piston on the gas is W_{CD}.
4. In the final process D → A (fig. 17.7 d), the base of the cylinder is replaced by a non-conducting wall and the gas is compressed adiabatically by increasing the load on the piston. The temperature of the gas increases from T_2 to T_1 and volume decreases from V_4 to V_1 and work done by the piston on the gas is W_{DA}.

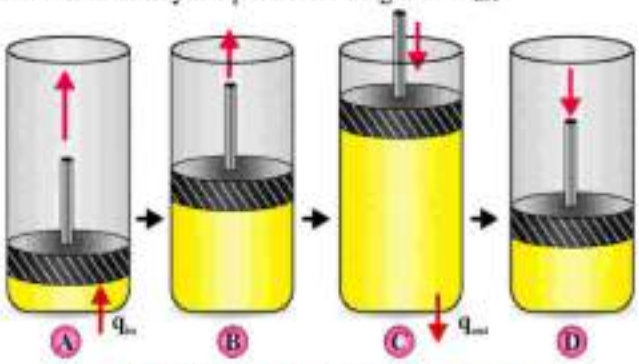


Fig: 17.7 (a,b,c,d) Carnot Engine cycle

If Q_1 is the heat absorbed by the working substance from the body in the isothermal expansion AB and Q_2 is the heat rejected to the cold body during the isothermal compression CD. In this cycle, the system comes back to its initial state and hence, there is no change in its internal energy. W is the external work done by the engine in one cycle; then from the 1st law of thermodynamics.

$$W = Q_1 - Q_2 \ldots\ldots\ldots\ldots\ldots 17.3$$

17.3.2 Efficiency of Carnot Engine:

Thermal efficiency (η) of a heat engine is the ratio of output to input.

$$Efficiency\ (\eta) = \frac{output}{input} = \frac{W}{Q_1} \ldots\ldots\ldots\ldots\ldots 17.4$$

From eq: 17.3, $W = Q_1 - Q_2$ then

$$\eta = \frac{Q_1 - Q_2}{Q_1} = 1 - \frac{Q_2}{Q_1} \ldots\ldots\ldots\ldots 17.5$$

Eq: 17.5 show that the efficiency of the engine increases as the ratio $\frac{Q_2}{Q_1}$ decreases. It can also be proved that the heat transferred to or from a Carnot engine is directly proportional to the temperature of the hot or cold body.

$$\frac{Q_2}{Q_1} = \frac{T_2}{T_1} \ldots\ldots\ldots\ldots 17.6$$

Thus, the efficiency of Carnot engine can be written as

$$\eta = 1 - \frac{T_2}{T_1} \ldots\ldots\ldots\ldots 17.7$$

The efficiency is usually taken in percentage, in that case.

$$\text{Percentage efficiency } (\eta) = \left(1 - \frac{T_2}{T_1}\right) \times 100\%$$

Thus, the efficiency of Carnot engine depends on the temperature of hot and cold reservoirs. The larger the temperature difference of two reservoirs the greater is the efficiency but it can never be one or 100% unless cold reservoir is at absolute zero temperature ($T_2 = 0^0K$)

Such reservoirs are not available and hence the maximum efficiency is always less than one. Nevertheless, the Carnot cycle establishes are upper limits on the efficiency of all heat engines.

No Practical heat engine can be perfectly reversible and also energy dissipation is inevitable. This fact is stated Carnot's theorem.

"No heat engine can be more efficient than a Carnot engine operating between the same two temperatures".

The Carnot's theorem can be extended to state that "all Carnot's engines operating between the same two temperatures have the same efficiency, irrespective of the nature of working substance". All real engines are less efficient than Carnot engine due to friction and other heat losses.

Worked Example 17.1

A Carnot's engine operates between two temperature reservoirs' of 227°C and 27°C, what is maximum efficiency of Carnot's engine?

Solution:

Step 1: Write down the known quantities and quantities to be found.

$T_1 = 227$ °C $T_1 = 227 + 273 = 500K$ $T_2 = 27$ °C $T_2 = 27 + 273 = 300K$

Maximum efficiency = η =?

Step 2: Write down the formula and rearrange if necessary.

$\eta = 1 - \frac{T_2}{T_1}$

Step 3: Put the values and calculate

$$\eta = 1 - \frac{300}{500} = 1 - 0.6 = 0.4$$

$\eta = 0.4 \times 100\%$

$\eta = 40\ \%$

Result: 40 %

17.4.1 Refrigerator:

According to the 2nd law of the thermodynamics heat will always flow spontaneously from hot to cold body, and never the other way around.

A refrigerator causes heat to flow from cold to hot by doing work, which cools the space inside the refrigerator.

In a heat engine, the direction of energy transfer is from the hot reservoir to the cold reservoir, which is the natural direction. The role of heat engine is to process the energy from the hot reservoir so as to do useful work. What we wanted to transfer energy from the cold reservoir to the hot reservoir, because that is not the natural direction of energy transfer. We must put some energy into a device to be successful. Devices that perform this task are called refrigerator.

For example, houses in summer are cooled using refrigerators called air conditioners. The air conditioner transfers energy from the cool room in the home to the warm air outside.

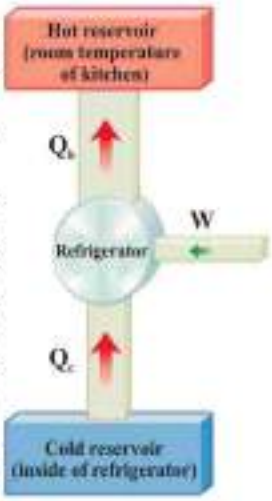


Fig: 17.8
Refrigerator or heat pump

In a refrigerator or heat pump, the engine takes in energy $|Q_C|$ from cold reservoir and expels energy $|Q_h|$ to a hot reservoir as shown in fig: 17.8, which can be accomplished only if work is done on the engine. From the first law of thermodynamics, we know that the energy given up to the hot reservoir must equal to sum of the work done and the energy taken in the form of the work done and the energy taken in the form of the cold reservoir,

$$Q_h = W + Q_c \ldots\ldots\ldots\ldots\ldots 17.8$$

Therefore, the refrigerator transfers energy from a colder body to a hotter body.

In Practice, a refrigerator includes a circulating fluid that gases through two sets of metal coils that can exchange energy with the surroundings. The fluid is cold and at low pressure when it is in the coils located in a cool environment, where it absorbs energy by heat. The resulting warm fluid is then compressed and enters the other coils as a hot, high-pressure fluid. There it releases its stored energy to the warm surroundings. In an air conditioner, energy is absorbed into the fluid in coils located in a building's interior; after the fluid is compressed, energy leaves the fluid through coil. In a refrigerator, the external coils are behind or underneath the unit Fig: 17.9 the internal coils are in the walls of the refrigerator and absorb energy from the food.

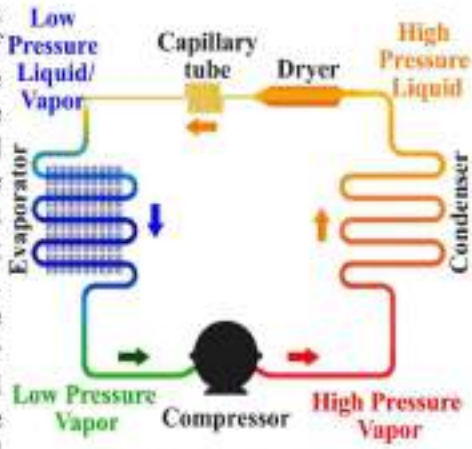


Figure 17.9 Refrigeration cycle

Efficiency of a Refrigerator:

The effectiveness of a refrigerator is described in terms of a number called the coefficient of performance (COP).

The COP is similar to the thermal efficiency for a heat engine in that it is a ratio of what you gain (energy transferred to or from a reservoir) to what you give (work input). For a refrigerator operating in the cooling mode, "what you gain" energy is removed from the cold reservoir the most effective refrigeration or air conditioner is one that removes the greatest amount of energy from the cooler reservoir in exchanges for the least amount of work. Therefore, for these devices operating in the cooling mode, we define the COP in terms of $|Q_C|$. The Coefficient of performance of a refrigerator is defined as

"The heat removed from the cold reservoir Q_C (i.e. inside refrigerator) divided by the work done to remove the heat (i.e. the work done by the compressor)".

$$\text{COP of refrigerator} = \frac{heat\ extracted}{work}$$

$$\text{COP} = \text{K} = \frac{Q_c}{w}$$

$$\therefore Q_h = Q_c + w$$

$$w = Q_h - Q_c$$

$$COP = K = \frac{Q_c}{Q_h - Q_c} \ldots\ldots\ldots\ldots 17.9$$

We know that for Carnot cycle $Q_1 \propto T_1$ and $Q_2 \propto T_2$

$$\therefore COP = K = \frac{T_c}{T_h - T_c} \ldots\ldots\ldots\ldots 17.10$$

T_c is the cryogenic temperature at which the heat is removed

T_h is the temperature at which the heat is rejected

It can also be seen from above relations that the coefficient of performance of refrigerator can be larger than 100% unlike the efficiency of heat engine which is always less than 100%. A good refrigerator should have a high COP, typically 5 or 6.

Worked Example 17.2

A certain refrigerator has a COP of 5.00. When the refrigerator is running, its power input is 500 watts. A sample of water of mass 500g and temperature 20.0°C is placed in the freezer component. How long does it take to freeze the water to ice at 0°C? Assume all other parts of refrigerator stay at the same temperature and there is no leakage of energy from the exterior, so the operation of the refrigerator result only is energy being extracted from the water.

Solution:

Step 1: Write down the known quantities and quantities to be found.

COP = K = 5.00,

P = 500 watt

Mass of water = m = 500g = 0.50 kg,

T_1 = 20°C

T_2 = 0°C,

$\Delta T = T_2 - T_1$ = -20°C

Specific heat of water = c = 4186 J/kg.°C

Latent heat of fusion $= L_f = 3.36 \times 10^5$ J/kg

Time = t =?

Step 2: Write down the formula and rearrange if necessary.

$$|Qc| = mc\Delta T - mL_f = m(c\Delta T - L_f)$$

$$= 0.50\,kg\,[(4186 J/kg.°C)(-20°C) - 3.36 \times 10^5\,J/kg]$$

$$= 2.08 \times 10^5 J$$

$$COP = \frac{|Q_c|}{w} \rightarrow w = \frac{|Q_c|}{COP} = \frac{2.08 \times 10^5}{5.00}$$

$$W = 4.17 \times 10^4 J$$

$$P = \frac{w}{t} \rightarrow t = \frac{w}{P} = \frac{4.17 \times 10^4\,J}{500w} = 83.4\,sec$$

Result: In reality the time interval for the water to freeze in a refrigerator is much longer than 83.4 s, which suggests that the assumptions of our model not valid. Only a small part of the energy extracted from the refrigerator interior in a given time interval comes from the water. Energy must also be extracted from the container in which the water is placed, and energy that continuously leak into the interior from the exterior must be extracted.

Worked Example 17.3

What is the coefficient of performance (COP) of a refrigerator that operates with Carnot efficiency between temperatures -3.00°C and +27.0°C?

Solution:

Step 1: Write down the known quantities and quantities to be found.

Tc = -3.00°C = 270K

Th = 27.0°C = 300K

COP = K =?

Step 2: Write down the formula and rearrange if necessary.

$$COP = K = \frac{Tc}{Th - Tc}$$

$$COP = K = \frac{270}{300-270} = \frac{270}{30} = 9$$

Result: The coefficient of performance of a refrigerator is 9.

Self-Assessment Questions:

1. An inventor claims that he has construct an engine which give efficiency of 56%. When working between -10 and 900, check whether his claim is true or false.
2. Why is the coefficient of performance (COP) used instead of efficiency in refrigerators?

17.5.1 Entropy:

Entropy was first expressed by German physicist and mathematician, Rudolf Clausius in 1865 into the study of thermodynamics to give a quantitative basis for the 2nd law of thermodynamics. The disorderness of the system is known as entropy. Entropy is also defined as

"The measure of a system's thermal energy per unit temperature that is unavailable for doing useful work".

We have discussed several processes that proceed naturally in the direction of increasing disorder. Irreversible heat flow increases disorder because the molecules are initially sorted into hotter and cooler regions; this sorting is lost when the system comes to thermal equilibrium. Adding heat to a body increases its disorder because it increases average molecular speed and therefore the randomness of molecular motion. Free expansion of a gas increases its disorder because the molecules have more randomness of position after the expansion than before.

Entropy provides a quantitative measure of disorder. T_0 introduce the concept, let's consider an isothermal expansion of an ideal gas; we add heat Q and let the gas expand just enough that the temperature remains constant. Because the internal energy of an ideal gas depends only on its temperature, the internal energy is also constant, so from the first law of thermodynamics the work w done by the gas during this expansion is equal to the heat Q added to it. That is

$$\Delta Q = \Delta W = P.\Delta V$$

The gas is in a more disordered state after the expansion than before, because the molecules are moving in a large volume and have more randomness of position. Thus ΔV is a measure of the increase in disorder of the molecules, and we see that it is proportional to the quantity $\frac{\Delta Q}{T}$. We introduce the symbol S, called the entropy of the system, and we define the entropychange ΔS during a reversible isothermal process as

$$\Delta S = \frac{\Delta Q}{T} \ldots\ldots\ldots\ldots 17.11$$

Change in entropy is positive when heat is added and negative when heat is removed from the system. Suppose an amount of heat Q flows from a reservoir at temperature T_1through a conducting rod to a reservoir at temperature when $T_1 > T_2$. The change in entropy of the reservoir at temperature T_1, which loses heat, decreases by $\frac{Q}{T_1}$, and of the reservoir at temperature T_2 which gains heat increases by $\frac{Q}{T_2}$. As $T_1 > T_2$ so $\frac{Q}{T_2}$ will be greater than $\frac{Q}{T_1}$. i.e. $\frac{Q}{T_2} > \frac{Q}{T_1}$. Hence, net change in entropy = $\Delta S = \frac{Q}{T_2} - \frac{Q}{T_1}$ is positive. It follows that in all natural processes where heat flows from one system to another; there is always a net increase in entropy. This is another statement of 2nd law of thermodynamics.

"The entropy of the universe during any process either remains constant or increases"

Similarly, we can prove that in every natural process the entropy always increases provided we considered the whole system. It seems to be a law of nature that a natural process always takes place in such a direction as to cause an increase in the entropy of the system and its surroundings. It is known as the law of increase of entropy.

The S.I unit of entropy is JK^{-1}. If heat is removed from a system, the change in entropy is written as negative.

17.5.2 Increase in Entropy Means Degradation of Energy:

Suppose a quantity of heat Q in a reservoir at temperature T_1. Let the temperature of the coldest available reservoir be T_o. A Carnot engine working between the temperatures T_1 and T_o can absorb heat Q at temperature T_1 and do useful work W_1 given by:

$$W_1 = Q - \theta_2$$

The efficiency of Carnot engine is

$$\eta = \frac{W_1}{Q} = 1 - \frac{T_o}{T_1}$$

$$W_1 = Q\,(1 - \frac{T_o}{T_1}) \ldots\ldots\ldots : 17.12$$

Equation (17.12) gives maximum available energy which can be converted into useful work, when heat Q is stored in the reservoir at temperature T_1.

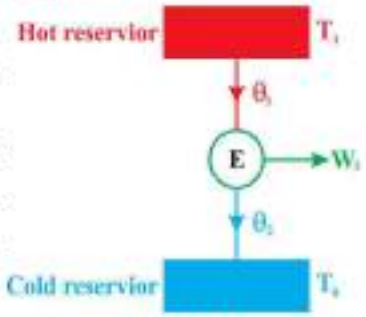


Figure:17.10 (a) Entropy

Consider an irreversible process, in which heat Q flows from the reservoir at temperature T_1, to another reservoir at a lower temperature T_2. A Carnot engine working between the temperatures T_2, and To can now take heat Q from the reservoir at temperature T_2, and do useful work W_2given by:

$$Efficiency = \eta = \frac{W_2}{Q} = 1 - \frac{T_o}{T_2}$$

$$W_2 = Q\,(1 - \frac{T_o}{T_2}) \ldots\ldots\ldots 17.13$$

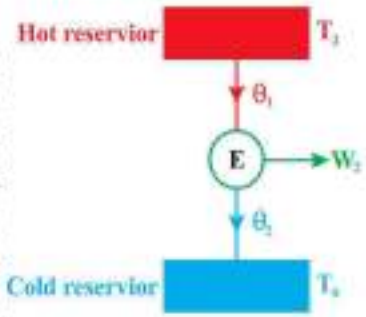


Figure:17.10 (b) Entropy

Equation (17.13) gives the maximum available energy which can be converted into useful work. When heat Q is stored in the reservoir at lower temperature T_2. As $T_2 < T_1$, we see from equation (17.12) and (17.13) that W_2 is less than W_1, i.e. available energy decreases with the increase of entropy during an irreversible process. Since all natural processes are irreversible, we conclude that the energy of the universe is continuously becoming unavailable for useful work. This is called the degradation of energy. From the equation (17.12) and (17.13), we get

$$W_1 - W_2 = Q\,(\frac{T_o}{T_2} - \frac{T_o}{T_1})$$

$$W_1 - W_2 = \left(\frac{Q}{T_2} - \frac{Q}{T_1}\right) T_o \ \ldots\ldots\ldots 17.14$$

Now $\frac{Q}{T_1}$ = decrease in entropy of the reservoir at temperature T_1 and $\frac{Q}{T_2}$ = increase in entropy of the reservoir at temperature T_2.

Thus $(\frac{Q}{T_2} - \frac{Q}{T_1})$ is increase in entropy of the universe. Also $(W_1 - W_2)$ is the amount of energy which has been degraded or made available for useful work. Hence equation (17.14)

shows that the increase of unavailable energy is equal to the increase in entropy of the universe multiplied by the temperature of the coldest available reservoir. Using eq: (17.14) an entropy increases, so unavailable energy also increases, so the useful energy decreases. It is called degradation of energy.

17.5.3 Energy is Degraded During All-Natural Processes:

It is believed that the temperature of the universe is increasing gradually and its entropy is also increasing. The mode of increase of entropy indicates that after a very long time, the entropy of the universe will be maximum and all the objects will be at the same temperature. At that time, no useful energy will be available and life will cease to exist, which is known as the "Heat Death" of the universe, which may not occur in near future.

It seems to be a law of nature that all natural processes always take place in such a direction so as to cause an increase in the entropy of a system and its surroundings.

Thus the 2nd law of thermodynamics is also expressed as the "Law of increase of entropy".

It states that, ***"when all the systems taking part in a process are included, the entropy either remains constant or increases".***

Symbolically, $\Delta S \geq 0$

The 2nd law of thermodynamics reveals that all the natural processes proceed towards a state of greater entropy, i.e. greater disorder or greater unavailability.

In all natural processes, energy always tends to pass from a more available (useful) form to a less available (degradation) form unavailability is a measure of entropy. A reversible process does not change total entropy of the universe. Thus $\Delta S = 0$; any irreversible process increases the total entropy of the universe $\Delta S > 0$, Total entropy change in one cycle of any Carnot engine is zero. All natural processes are irreversible and therefore, during all-natural process's energy is degraded.

17.5.3 Systems Tends to Become Less Orderly Over Time:

The disorder of the system increases during any natural process, a stage will reach when universe will approach a stage of maximum disorder. At this stage, matter will become a uniform mixture i.e. whole universe will be at uniform temperature. No work can then be done and heat flow will cease.

Time arrow concept of entropy tells us in which direction the time is going. It is the change in entropy that ultimately provides us with the answer to why systems will naturally evolve in one direction with time and not the other. Systems always evolve in time in such a way that the total entropy of system plus environment increases. If you observe a system in which the entropy appears to decrease, you can be sure that somewhere there is change in the entropy of the environment large enough to make the total entropy change positive. Entropy has been called a time arrow: i.e. events occur in the direction of increasing disorder with time.

In all the processes going on in practical life, the entropy always increases with time and disorder is produced more and more. Hence this increase of entropy is identification for the time passage.

Worked Example 17.4

What is the change in entropy of 30gm of water at 0°C as it is changed into ice at 0°C? Take latent heat of fusion of ice 336000 J kg^{-1}.

Solution:

Step 1: Write down the known quantities and quantities to be found.

Change in entropy = ΔS =?

Mass of water = m =30gm = 0.03kg

T = 0°C = 0 + 273 = 273K

Specific latent heat of fusion of ice = H_f = 336000J kg^{-1}

Step 2: Write down the formula and rearrange if necessary.

$\Delta S = \frac{\Delta Q}{T}$

∴ Heat required freezing 30gm of water at 0°C to ice at 0°C.

$\Delta Q = mH_f = 0.03 \times 336000 = 10800J$

Put the values and calculate

$\Delta S = \frac{10080\ J}{273\ K} = 36.92\ J\ k^{-1}$

Result: The change in entropy comes out 36.92 Jk^{-1}.

Worked Example 17.5

Calculate change in entropy when 10kg of water is heated from 90°C to 100°C. Specific heat of water = 4180J.$kg^{-1}K^{-1}$.

Solution:

Step 1: Write down the known quantities and quantities to be found.

Change in entropy = ΔS =?

Mass of water = m =10kg

T_1 = 90°C

T_2 = 100°C

ΔT = 100 – 90 = 10°C

Average temperature = $T = \frac{T_1+T_2}{2} = \frac{90+100}{2}$ =95°C

T = 95°C + 273 = 368K

Step 2: Write down the formula and rearrange if necessary.

$\Delta S = \frac{\Delta Q}{T} = \frac{mc\Delta T}{T}$

Put the values and calculate

$\Delta S = \frac{10 \times 4180 \times 10}{368} = 1135.87\ Jk^{-1}$

Result: The change in entropy comes out 1135.87 Jk^{-1}

SUMMARY

- Second law of thermodynamics can be states as:
  - There is no perpetual motion machine that can convert the given amount of heat completely into work.
  - The total entropy of any system plus that of its environment increases as a result of any natural process.
- A heat engine is a device which converts a part of thermal energy into useful work.
- Reversible Process: A process in which system and its surroundings can be returned to their initial conditions.
- Irreversible Process: A process which cannot be retraced in the backward direction by reversing the controlling factors.
- Efficiency of Carnot Engine is the ratio of output to input temperature and is $1 - \frac{T_2}{T_1}$
- Refrigerator is a heat engine operating in reverse as that of an ideal engine.
- Coefficient of performance of refrigerator (COP) is equal to the ratio of the heat taken from the cold reservoir to the work done by the refrigerator.

$$COP = \frac{T_c}{T_h - T_c}$$

- T_c is the cryogenic temperature at which the heat is removed and T_h is the temperature at which the heat is rejected.
- Entropy:
  - It is a measure of disorder (microscopic randomness) of a system.
  - It is a measure of unavailable energy of a system.
  - In terms of 2nd law of thermodynamics, "Every process in nature takes place such that the entropy of an isolated system either remains unchanged or it increases". $\Delta S \geq 0$.
- Entropy change ΔS due to heat transfer at absolute temperature T is given by $\Delta S = \pm \frac{\Delta Q}{T}$
- The S.I units of entropy are $J.K^{-1}$

4. Which one of the following processes is irreversible?
   (a) Slow compression of an elastic spring
   (b) Slow evaporation of a substance in an isolated vessel
   (c) Slow compression of a gas
   (d) A chemical explosion
5. According to 2nd law of thermodynamics, 100% conversion of heat into mechanical work is:
   (a) Possible (b) Possible, if the conditions are ideal
   (c) Not possible (d) Possible, if the process is adiabatic
6. The change in entropy is given by
   (a) $\Delta S = \frac{\Delta Q}{T}$ (b) $\Delta S = \Delta Q.\ T$ (c) $\Delta S = \frac{\Delta U}{T}$ (d) $\Delta S = \frac{\Delta W}{T}$
7. The net change in entropy as a system in a natural process is
   (a) Positive (b) Negative (c) Zero (d) Infinite
8. The efficiency of diesel engine is
   (a) Greater than petrol engine (b) Less than petrol engine
   (c) Equal to petrol engine (d) both have efficiency 1
9. Second law of thermodynamics states that:
   (a) Energy can't be converted. (b) Entropy decreases over time.
   (c) Heat flows from cold to hot. (d) Entropy increases over time
10. The process violates the 2nd law of thermodynamics is:
   (a) Refrigerator cooling. (b) Heat engine working.
   (c) Gases mixing. (d) Heat flowing from cold to hot.

Section (B): CRQs (Short Answered Questions):

1. What are some factors that affect the efficiency of automobile engines?
2. What happens to the temperature of a room in which an air conditioner is left running on table in the middle of the room?
3. Under what conditions can heat be added to a system without changing its temperature?
4. Is it possible to cool a room by keeping the refrigerator door open?
5. When does the entropy of a system decrease?
6. Is it possible, according to the 2nd law of thermodynamics to construct an engine that is free from thermal pollution?
7. Can heat be completly converted to work?
8. Define that why entropy has often been called as "time arrow".

Section (C): ERQs (Long Answered Questions):

1. Give the two statements of the second law of the thermodynamics. Elaborate the concept of entropy and state the second law of thermodynamics in terms of this concept.
2. Explain the working principle of heat engine and also derive the formula for its efficiency.
3. Describe the concept of reversible and irreversible process.
4. What is Carnot engine? Give the operation of Carnot cycle and show that the efficiency of even this engine is less than 100%.

5. Describe that refrigerator is a heat engine operating in reverse as that of an ideal heat engine and find its efficiency.
6. Explain that change in entropy is positive when heat is added and negative when heat is removed from the system.
7. Explain that increase in entropy means degradation of energy.

Section (D): Numerical:

1. A Carnot engine takes 2000J of heat from a reservoir at 500K does some work, and discards some heat to a reservoir at 350K. How much heat is discarded, how much work does the engine do, and what is the efficiency? **(Ans: 1400J, 600J, 30%)**
2. One kilogram of ice at 0°C is melted and converted to water at 0°C. Compute its change in entropy. **(Ans: 1220 $J.K^{-1}$)**
3. In a high-pressure steam turbine engine, the steam is heated to 600°C and exhausted at about 90°C. What is the highest possible efficiency of any engine that operates between these two temperatures? **(Ans: 58.4 %)**
4. Temperature difference between the surface water and bottom water in Manchester Lake might be 5°C.Assuming the surface water to be at 20°C. What highest efficiency a steam engine could have if it operates between these two temperatures? **(Ans: 1.71%)**
5. A heat engine works at the rate of 500kW. The efficiency of the engine is 30%. Calculate the loss of heat per hour. **(Ans: 4.2×10^9 J)**
6. A heat engine performs work of 0.4166 watts in one hour and rejects 4500J of heat to the sink. What is the efficiency of engine? **(Ans: 24.9%)**
7. A Carnot engine operates between the temperatures 850K and 300K. the engine performs 1200J of work in each cycle, which takes 0.25 sec
   (a) What is the efficiency of this engine?
   (b) What is the average power of this engine?
   (c) How much energy is extracted as heat from the high temperature reservoir?
   (d) How much energy is delivered as heat to the low temperature reservoir?
   [Ans (a) 65% (b) 4.8KW (c) 1855J (d) 655J]
8. A Carnot engine absorbs 52kJ as heat and exhaust 36kJ as heat in each cycle. Calculate:
   (a) The engine efficiency
   (b) The work done per cycle in kilojoules. **[Ans (a) 30.76% (b) 16KJ]**

Introduction:

A magnetic field is a fundamental concept in physics that describes the region around a magnet or a current-carrying conductor where magnetic forces are exerted on other magnets, conductors, or charged particles.

Permanent magnets, such as iron magnets, have their own intrinsic magnetic fields due to the alignment of their magnetic domains.

Magnetic fields have numerous practical applications, including:

- Electric motors
- Transformers
- MRI machines (Magnetic Resonance Imaging)
- Magnetic levitation trains
- Magnetic compasses
- Magnetic data storage (hard disk drives)

18.1 Magnetic Fields:

Magnetic field can be defined as ***"An area around a permanent magnet or a current carrying conductor, where they generate a magnetic force".*** Unlike in electrostatics, where electric charges create an electric field and exert electric forces on other charges, the magnetic field lacks a direct counterpart due to the absence of magnetic monopoles.

Magnetic fields can be generated through two primary methods:

- From permanent magnets.
- From current carrying conductors, also known as electromagnets

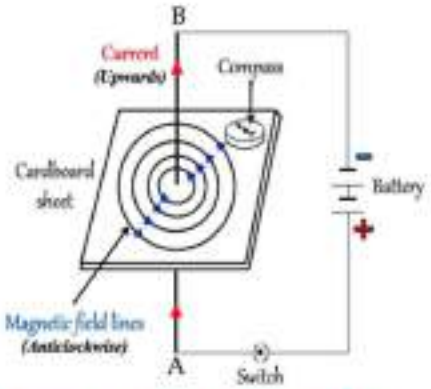


Figure 18. 1 (a) Current Carrying Wire

In 1819, Hans Christian Oersted made a significant discovery that revealed how a current Carrying conductor generates a magnetic field. The experiment depicted in Figure 18.1(a) clearly demonstrates this phenomenon. Simple setup consisting of a wire (AB) passing through a flat cardboard and connected to a battery via a switch. When the switch is closed, completing the circuit and enabling current flow through the wire, a circular magnetic field is generated around it. When iron fillings are scattered on to the cardboard, they become magnetized and arrange themselves in a circular pattern around the current-carrying wire. Additionally, a magnetic compass placed near wire AB indicates the direction of the magnetic field.

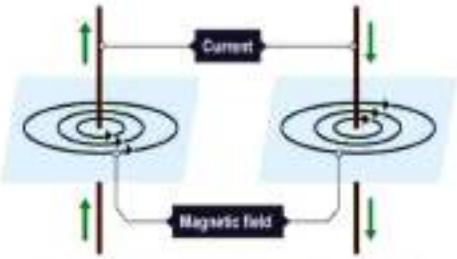


Figure 18.1 (b) Clockwise and Anticlockwise direction of magnetic fields

These observations illustrate that the magnetic field's direction produced by the current carrying wire follows the right-hand rule. According to this rule, when you grasp the wire with your right hand, your thumb points in the direction of the current, while the curl of your fingers indicates the direction of the magnetic field line. If the direction of the current is reversed, the needles also reverse their orientation. Conversely, if the current's direction is reversed, it generates a magnetic field in a clockwise direction. As shown in figure 18.1 (b).

18.2 Magnetic Force on a current carrying conductor:

Electric current is the orderly flow of charges through a conductor. As electrons move, they create a magnetic field around them, in addition to the electric field they produce. When it comes to a current-carrying conductor, the electric field generated by the moving electrons is counter balanced by the electric field produced by stationary protons within the conductor. Consequently, a magnetic field is produced around the current-carrying conductor.

When we place a current-carrying conductor with in a uniform External magnetic field, the interaction between the magnetic field produced by the conductor and the external magnetic field gives rise to external force, denoted as F, acting on the conductor, as illustrated in figure18.2.

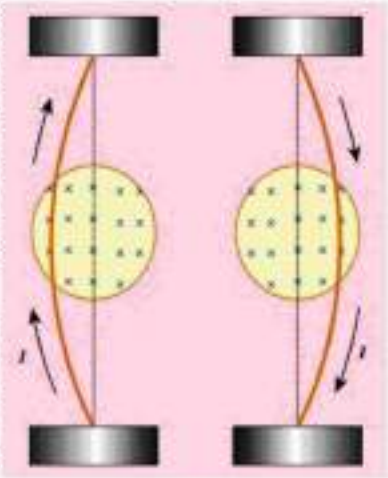


Figure 18.2 Magnetic Force on a current carrying conductor

A vertical wire is securely anchored at both ends and passes through a gap in a magnet. The external magnet generates a uniform magnetic field directed into the page. When a current flows upward through the wire, the wire deflects to the left. Conversely, reversing the current causes the wire to deflect to the right.

18.2.1 Factors on which the force acting on current carrying conductor in a magnetic field:

The force acting on a current-carrying conductor depends on several factors, such as, length, current, and the strength of the external magnetic field

The force (F) is directly proportional to the length of the conductor (L) that lies within the magnetic field

$$F \propto L$$

The force is also directly proportional to the current (I) passing through the conductor.

$$F \propto I$$

Similarly, the force is directly proportional to the strength of the external magnetic field (B).

$$F \propto B$$

Combining these factors, we will obtain the following relationship:

$$F \propto BIL$$

Here, 'k' is the proportionality constant, and in the SI unit system, its value is equal to 1.
Therefore, we can express the force as: $F = BIL$..... (18.1)
If the magnetic field is not perpendicular to the wire, Equation (18.1) can be modified to:

$F = I\,(L \times B)$ where 'X' denotes the vector cross product.

The magnitude of the force **(F)** is determined by: $F = BILSin\theta$
Where, θ represents the angle between the direction of the conductor's length ($\boldsymbol{L}$) and the external magnetic field ($\boldsymbol{B}$).

Maximum Force:

The deflecting force reaches its maximum when the angle (θ) between the length of the conductor and the external magnetic field is θ = 90 degrees, assuming other factors remain constant.

Minimum Force:

Conversely, the deflecting force is minimized when θ=0 degrees, signifying that the magnetic field and conductor are aligned.
Since magnetic force is a vector quantity, its direction can be determined using Fleming's left-hand rule, which provides a straight forward way to establish the direction of the force in relation to the current and magnetic field.

Fleming's left-hand rule:

Fleming's left-hand rule is used to determine the direction of force acting on a current carrying wire placed in a magnetic field, also to identify the direction of the forces in an electric motor in order to apply Fleming's left-hand rule, follow these steps: Extends our fore finger, thumb, and Middle finger such that they are mutually perpendicular as shown in figure 18.3.

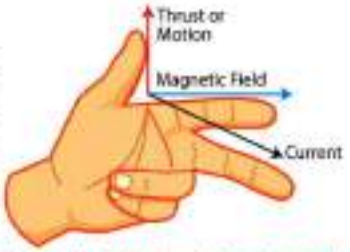


Figure 18. 3 Fleming's left-hand rule

Worked Example 18.1

A current carrying conductor of certain length placed at 90° to the magnetic field experiences a force F. What will be the force if the current is increased four times, length is halved and magnetic field is tripled?

Solution:

Step1: Write down the known quantities and quantities to be found.

$I = 4I$, $L = L/2$, $B = 3B$ $F' = ?$

Step2: Write down the formula and rearrange if necessary.

As we know that force on current carrying wire placed in magnetic field is

$$F = BIL\text{Sin}\theta$$

When current is increased four times, length is halved and magnetic field tripled then,
Force,

$$F' = (4I)\ x\ \left(\frac{l}{2}\right)\ x\ (3B)\ F = 6\,F.$$

Result: Therefore, the force increases six times.

Worked Example 18.2

A conductor of length 50cm carrying a current of 5A is placed perpendicular to a magnetic field of induction 2×10^{-3} T. Find the force on the conductor.

Solution:

Step1: Write down the known quantities and quantities to be found.

$I = 5A$, $L = 50\times10^{-2}$m, $B = 2\times10^{-3}$T, $F = ?$

Step2: Write down the formula and rearrange if necessary.

Force on the conductor

$$F = BIL = 5 \times 50 \times 10^{-2} \times 2 \times 10^{-3}$$
$$F = BIL = 5 \times 10^{-3}\ N$$

Result: Therefore, $F = 5\times10^{-3}$ N

18.2.2 Magnetic Flux:

Magnetic flux is defined as **the number of magnetic lines of force passing through an area held perpendicular to it**. It is calculated by taking the dot product between the magnetic field and area vector. The dot product ensures that magnetic lines must pass perpendicular to the given area. Experimentally magnetic flux is usually measured with a Flux meter. Magnetic flux through an area A placed in uniform magnetic field (Figure18.4 is calculated by using formula as given below

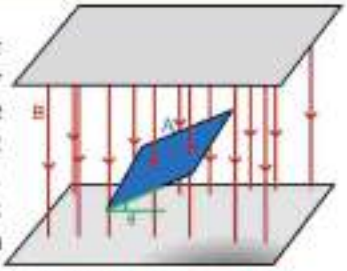


Figure 18.4 Magnetic Flux

$\emptyset_B = B.A$ **or** $\emptyset_B = BA\cos\theta$ ...(18.2)

Where B is uniform magnetic field, **A** is an area vector whose magnitude is equal to the surface element and directed perpendicular to the given surface element.

Maximum Magnetic Flux:

The magnetic flux will be maximum when the angle between magnetic field **B** and area vector **A** is zero, i.e. $=0^0$ and surface will be perpendicular to the magnetic lines. This case is shown in figure 18.5(a).

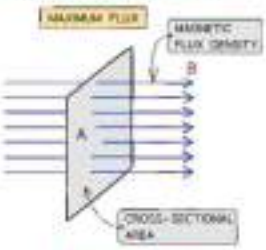


Figure 18.5 (a)
Maximum Magnetic Flux

Minimum Magnetic Flux:

Similarly when the angle between area vector and magnetic field is 90^0 then lines of force donot pass through the surface and magnetic flux will be minimum as shown in figure18.5 (b).

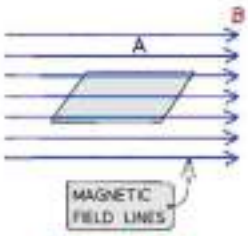


Figure 18.5 (b)
Minimum Magnetic Flux

Unit of Magnetic Flux:

The SI unit of magnetic flux is Weber (Wb) =

$$1Wb = \frac{Nm}{A}$$

It is named after renowned German Physicist and co-inventor of telegraph.

$\emptyset_B = \boldsymbol{B.A}$, represents the magnetic flux $\emptyset_B$ through a surface (**A**) in a magnetic field (**B**). To solve problems using this formula, you will typically need to work with scenarios where you know the magnetic field strength(B) and the area(A) through which the magnetic field lines pass.

Worked Example 18.3

Calculate the magnetic flux through a rectangular loop with dimensions 5 cm × 10 cm placed in a magnetic field of strength 0.2 T. The loop is oriented such that the magnetic field lines are perpendicular to the plane of the loop.

Solution:

Step1: Write down the known quantities and quantities to be found.

Magnetic field strength (B)=0.2T

$$Area\ of\ the\ loop\ (A) = 0.05\ m \times 0.10m = 0.005\ m^2$$

Step2: Write down the formula and rearrange if necessary.

$$\emptyset_B = BAcos\theta$$

Step 3: Put the values and calculate

$$\emptyset_B = BAcos\theta$$

$$\emptyset_B = \mathbf{0.2T \times 0.005m^2 = 0.001\ Wb\ (Weber)}$$

Result: So, the magnetic flux through the rectangular loop is 0.001 Weber.

DO YOU KNOW?

The magnetic properties of the disk's platters and the read/write head's ability to generate and detect magnetic fields play a crucial role in storing, encoding, and accessing data in computer systems.

18.2.3 Magnetic Flux Density:

Magnetic flux density is described in relation to the magnetic force experienced by a current carrying conductor placed perpendicularly to a magnetic field. For a uniform magnetic field, the flux density B is expressed by the equation: $B = \frac{F}{IL}$

Hence magnetic flux density can be defined as:

The magnetic flux density at a specific point in space is the force experienced per unit length by along straight conductor carrying unit current, placed perpendicular to the field at that particular location.

Magnetic flux density is most pronounced near the pole of a bar magnet and gradually decreases with distance from it.

Unit of Magnetic Flux Density:

The SI unit of magnetic flux density is Tesla (T), named after the Serbian-American inventor Nikola Tesla.

One Tesla: is defined as follows, If a conductor having length 1m and carries a current of 1 A placed perpendicularly to the magnetic field experience a force of one Newton then magnetic flux density will be 1Tesla.

In essence, the Tesla provides a standardized measure of the strength of a magnetic field, with larger values indicating stronger magnetic fields and smaller values representing weaker fields.

18.3 Measurement of Magnetic Flux density by current balance:

A simple experimental setup designed for the measurement of flux density between two magnets, As shown in figure 18.6. The magnetic field within this magnet configuration is uniform. To measure the length (L) of the current-carrying wire within this uniform magnetic field, a ruler can be used. When the wire carries no current, the magnet arrangement is positioned on top of balance, and the balance is calibrated to read zero. Subsequently, when an electric current (I) flows through the wire, the ammeter displays its magnitude. The wire experience an upward force, and in accordance with Newton's third law of motion, an equal and opposite force acts upon the magnets. Consequently, the magnets are pushed down wards, causing the balance to indicate a reading. The force (F) can be calculated as F = mg, with 'm' representing the mass indicated on the balance in kilograms and 'g' representing the acceleration due to gravity ($9.81 ms^{-2}$).

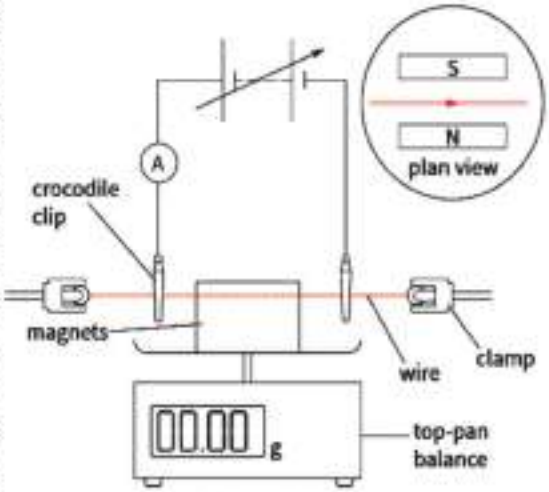


Figure 18.6
Experimental setup for the measurement of flux density

With the knowledge of F,I, and L, one can
Determine the magnetic flux density (B) between the magnets using the following equation

$$B = \frac{F}{IL} \ldots\ldots\ldots (18.3)$$

18.4 Amperes Law:

Ampere's Law is an important principle in electromagnetism. It deals with the behavior of electric currents and their associated magnetic fields. It is named after Andre-Marie Ampere, a French physicist and mathematician who made significant contributions to the study of electromagnet is mid the 19th century.

It is known that an electric current flowing through a wire creates a magnetic field in its vicinity. When we envision a closed circular path with the wire positioned at its center, the Magnetic flux density (B) within this circular region changes, depending on both the current (I) and the distance (r) from the wire as shown in figure18.7.

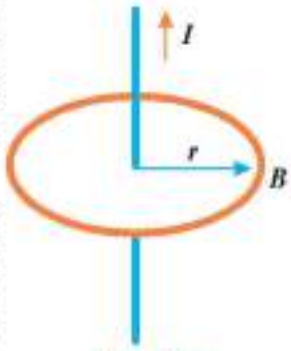


Figure 18.7
Current carrying wire

$$B \propto 2I \text{.....(i)}$$
$$\text{Or } B \propto \frac{1}{r} \text{.....(ii)}$$

Combing (i) and (ii), we get

$$B \propto \frac{2I}{r}$$
$$B = \frac{\mu_0 I}{2\pi r}$$

Where μ_0 is permeability of free space = $4\pi \times 10^{-7}$ Wb A^{-1} m^{-1}

Now consider an Amperian loop in form of a closed path around the current carrying wire as shown in figure 18.7. For any path element ΔL we can write

$$B.\Delta L = \mu_0 I$$

Since angle between B and L is zero so $B.\Delta L = B\Delta L = \mu_0 I$

Now adding up over whole closed path

$$\sum B.\Delta L = \mu_0 I \text{ ..... (18.4)}$$

Equation (18.4) is the mathematical expression of Ampere law. Hence Ampere's law can be defined as the integral of the magnetic field (**B**) around a closed loop is directly proportional to the total electric current (**I**) passing through the loop.

Application of Ampere's Law:

(i) Solenoid:

A solenoid is along, tightly wound coil of wire, and often used to generate a uniform magnetic field within the coil's interior. It consists of many closely spaced turns of wire, and when an electric current flow through it, it creates a magnetic field that is similar to the field produced by a bar magnet. To determine the magnitude of B of a solenoid let us Consider an Amperian loop abcd with lengths l_{ab},l_{bc}, $l_{cd}$and l_{da}as shown in figure18.8.

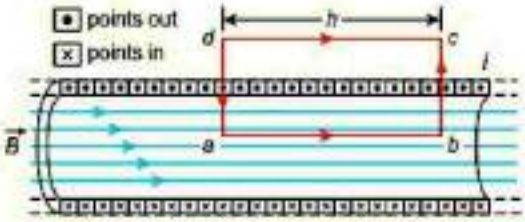


Figure 18.8 Amperian loop Solenoid

Then from Ampere's law

$$\sum B.\Delta l = \mu_0 I$$

The left hand side of equation is

$$\sum B.\Delta l = B.l_{ab} + B.l_{bc} + B.l_{dc} + B.l_{da}$$

As the element al lengths along bc and da are perpendicular to the magnetic field so

$$B.l_{bc} = 0 \; and \; B.l_{da} = 0$$

Moreover, the magnetic field intensity outside of solenoid ≈ 0, so $B.l_{cd}=0$
Hence the magnetic field due to solenoid is given by

$$\sum B.\Delta l = B.l_{ab} = \mu_0 I$$

For N number of turns on the solenoid

$$\sum B.\Delta l = B.l_{ab} = \mu_0 NI$$

The number of turns per unit length is given by $\boldsymbol{n} = \dfrac{\boldsymbol{N}}{\boldsymbol{l_{ab}}}$

$$B = \mu_0 nI$$

Hence it is shown that B is independent of the position within the solenoid which shows that the field is uniform with in a solenoid.

Worked Example 18.4

Suppose we have a solenoid with 500 turns per meter and a current of 2.5 amperes flowing through it. Calculate the magnetic flux density inside the solenoid.

Solution:

Step1: Write down the known quantities and quantities to be found.

The no of turns = 500I = 2.5A

$\mu_0 = 4\pi \times 10^{-7}$

Magnetic flux density inside the solenoid,

(B) = ?

Step2: Write down the formula and rearrange if necessary.

$$\boldsymbol{B = \mu_0 nI}$$

Step 3: Put the values and calculate.

$\boldsymbol{B = \mu_0 nI}$

$B = 4\pi \times 10^{-7} \times 500 \times 2.5$

$B = 3.14 \times 10^{-3}$ T B = 3.14 mT

Result: So, the magnetic flux density inside the solenoid is 3.14 milli-tesla.

(ii) Toroid:

A toroid can be considered as a solenoid that is bent into the shape of a circle. It also generates a magnetic field within its interior. The magnetic field lines inside a toroid are generally circular and concentrated within the coil.

Suppose a toroid consist of N closely packed turns of copper wire and carry a current I as shown in 18.9.

Imagine a hollow circular ring wound with N number of turns of current-carrying wire. In this case, within the toroid, point P is our point of interest, and we will denote the magnetic field at this location as B. Inside the toroid, we construct an Amperian loop forming a circle that passes through point P. This

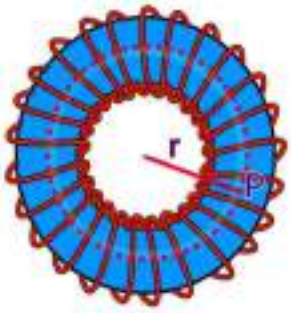


Figure 18.9 A toroid

arrangement gives rise to concentric circles within the toroid.
Because of the symmetrical nature of the magnetic field, the magnitude of the field at all points along the circle is uniform, and the field direction is tangential to the circle.
Therefore,

$$\sum B.\Delta l = \sum B\Delta l Cos(0) = B\sum \Delta l = (2\pi r)$$

Now, we can apply Ampere's law:

$$(2\pi r) = \mu_0 NI$$
$$B = \mu_0 \frac{NI}{2\pi r}$$

This is the expression for the magnetic field strength inside a toroid. It is inversely proportional to the radius r and directly proportional to the number of turns per unit length N and the current I flowing through the turns.

Worked Example 18.5

Suppose we have a toroid with 1000 turns and a current of 3 amperes flowing through it. The toroid has a circular cross-section with a radius of 0.1 meters. Calculate the magnetic flux density inside the toroid.

Solution:

Step 1: Write down the known quantities and quantities to befound.

The no of turns = 500I = 2.5A

$\mu_0 = 4\pi \times 10^{-7}$

Magnetic flux density inside the solenoid,

(B) = ?

Step 2: Write down the formula and rearrange if necessary.

$$B = \mu_0 \frac{NI}{2\pi r}$$

Step 3: Put the values and calculate

$$B = \mu_0 \frac{NI}{2\pi r}$$
$$B = (4\pi \times 10^{-7} \times 1000 \times 3)/(2\pi \times 0.1)$$
$$B = (4\pi \times 10^{-7} \times 1000 \times 3)T$$
$$B = 3.77 \times 10^{-3}T \text{ or } B = 3.77mT$$

Result: ***The magnetic flux density inside the toroid is 3.77milli-tesla.***

18.5 Force on a charged particle in a uniform magnetic field:

When a charged particle moves through a uniform magnetic field, it experiences a force known as the magnetic force. This phenomenon is described by the Lorentz force equation, which can be written as: $\vec{F} = (\vec{v} \times \vec{B})$.

Apparatus and Materials:

1. **Cathode Ray Tube (CRT):** This is a vacuum tube that produces a beam of electrons. It consists of an electron gun, focus in, deflection plates, and a fluorescent screen as shown in figure 18.10.
2. **Magnetic Field Source:** We'll need a strong and uniform magnetic field source, such as a Helmholtz coil or a solenoid.
3. **Voltage Source:** A variable voltage source to create an electric field.
4. **Fluorescent Screen:** A screen coated with a phosphorescent material to visualize the electron beam.
5. **Rulers and Measurement Devices:** To measure the radius of the electron beam's circular path and the electric potential applied.

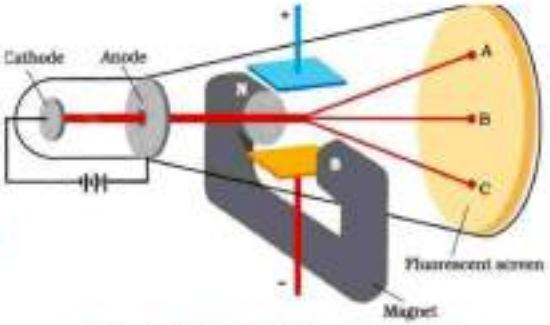


Figure 18.10 Cathode Ray Tube (CRT)

Experimental Procedure:

1. **Setup:**
   - Set up the CRT in a vacuum chamber to ensure the electron beam doesn't interact with air molecules.
   - Position the Helmholtz coil (or solenoid) around the CRT, providing a uniform magnetic field parallel to the beam path.
   - Connect the voltage source to the focusing and deflection plates to create an electric field perpendicular to the magnetic field.
2. **Calibration:**
   - Calibrate the magnetic field by measuring its strength at the location of the electron beam. We can do this using a magnetic field strength meter.
3. **Electron Beam Production:**
   - Apply a high voltage across the cathode and anode of the CRT to generate a beam of electrons from the cathode (electron gun).
   - Use the focusing and deflection plates to control and direct the electron beam.
4. **Observe the Electron Beam:**
   - Turn off the magnetic field and electric field to observe the electron beam's straight-line path on the fluorescent screen.
5. **Apply the Magnetic Field:**
   - Turn on the magnetic field, which causes the electron beam to bend in a circular path due to the Lorentz force (the interaction between the magnetic field and the moving electrons).
6. **Measure the Radius(r):**
   - Measure the radius of the circular path formed by the electron beam on the screen. Ensure the screen is marked with a scale for accurate measurement.

Construction:

1. **Coil**: The key component of a Galvano meter is a coil of wire (usually wound around as oft iron core) suspended within a magnetic field. The coil is mounted on a spindle so that it can rotate freely.
2. **Magnet**: A permanent magnet or an electromagnet is placed around the coil. The magnetic field lines from the magnet pass through the coil.
3. **Spring**: A delicate torsion spring is attached to the coil, providing a restoring torque when the coil is deflected.
4. **Pointer**: A thin pointer or needle is attached to the coil, allowing for the measurement of the deflection.

Figure 18.12 schematic diagram of Galvanometer

Working:

- When a small electric current flow through the coil, it generates a magnetic field around the coil according to Ampere's law. This magnetic field interacts with the external magnetic field (provided by the permanent magnet or electromagnet) to exert a torque on the coil.
- The torque causes the coil to rotate, and the amount of rotation is proportional to the current passing through it. This rotation is indicated by the deflection of the pointer on a calibrated scale.
- The coil continues to rotate until the restoring torque from the spring equals the torque due to the current-induced magnetic field. At this point, the pointer comes to rest, and its position on the scale indicates the magnitude of the current.
- Consider a rectangular coil consisting of N turns, each with a current I flowing through them and across-sectional area A. When this coil is situated within a uniform radial magnetic field B, it undergoes a torque as shown in figure 18.13.

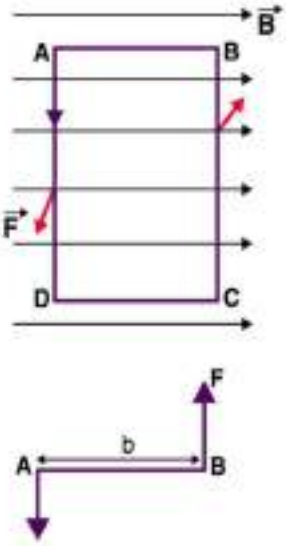


Figure 18.13 torque on rectangular coil

- Let's examine a single turn ABCD of the rectangular coil, characterized by a length L and breadth b as shown in figure 18.13. This turn is suspended within a magnetic field with strength of B, arranged so that the coil's plane is parallel to the magnetic field lines. As sides AB and DC are aligned parallel to the magnetic field, they donot experience any discernible force due to the magnetic field. However, sides AD and BC, which are perpendicular to the field's direction, encounter an effective force denoted as F, given by the equation F= BIL.
- By employing Fleming's left-hand rule, we can recognize that the forces acting on sides AD and BC are oriented in opposite directions to each other. When a pair of equal and

opposite forces, denoted as F and collectively referred to as a couple, act on the coil, they generate a torque. This torque induces a deflection in the coil.

- The torque(τ) is calculated as the product of the force(F) and the perpendicular distance (b) between these forces: $\tau = F \times b$
- Substituting the known value of F, we have:
  Torque(τ) acting on a single-loop ABCD of the $coil = BIL \times b$
  Where $L \times b$ represents the area (A) of the coil. Consequently, the torque acting on a coil with nturnsis given by: $\tau = NIAB$

This magnetic torque causes the coil to rotate, leading to the twisting of the phosphor bronze strip. Simultaneously, the spring (S) attached to the coil exerts a counter-torque, known as there storing torque (kθ), resulting in a stable angular deflection.

Under equilibrium conditions, we find: $k\theta = NIAB$

Here, k is termed the torsional constant of the spring, representing the restoring couple per unittwist. The deflection or twist (θ) is quantified as the reading displayed on a scale by a pointer connected to the suspension wire.

$$\theta = (NAB/k)I$$

Therefore $$\theta \propto I$$

The quantity (NAB/k) is a constant for a given galvanometer. Therefore, it is evident that the deflection observed in the galvanometer is directly proportional to the current flowing through it.

18.6.2 Conversion of Galvanometer to Ammeter:

To convert a galvanometer into an ammeter, a low-resistor known as a shunt is connected in parallel with the galvanometer. The selection of an appropriate shunt is based on the desired current range of the ammeter.

From the circuit as shown in figure 18.14.

- R_g represents the resistance of the galvanometer.
- G denotes the Galvanometer coil.
- I stand for the total current flowing through the circuit.
- I_g represents the total current passing through the galvanometer, which corresponds to the full-scale reading.
- R_s represent the value of the shunt resistance.

When the current I_g lows through theGalvano meter, the current passing through the shunt is determined by

$$I_s = I - I_g.$$

The voltages across both the galvanometer and shunt are equal due to their parallel connection.

Hence, we can establish the following equation:

$$R_g I_g = (I - I_g) \times R_s$$

$$R_s = \frac{R_g I_g}{I - I_g}$$

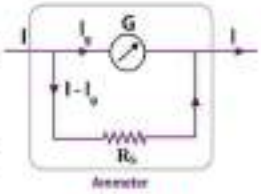


Figure 18. 14
Ammeter and its schematic diagram

Worked Example 18.6

A moving coil galvanometer of resistance 100Ω is used as an ammeter using a Shunt of 0.1Ω. The maximum deflection current in the galvanometer is 100μA. Find the current in the circuit so that the ammeter gives maximum deflection.

Solution:

Step1: Write down the known quantities and quantities to be found.

R_g=100Ω

Rs= 0.1Ω,

I_g=100μA

Step2: Write down the formula and rearrange if necessary.

$R_g \times I_g = (I - I_g) \times R_S$

Step 3: Put the values and calculate.

$$I = \frac{(R_g \times I_g + g \times Rs)}{R_S}$$

$$I = \frac{(1 + R_g)}{R_S} \times I_g$$

Result: We get $I = 100.1mA$

Conversion of Galvanometer to Voltmeter

A galvanometer is converted into a voltmeter by connecting it in series with high resistance as shown in figure 18.15. A suitable high resistance is chosen depending on the range of the voltmeter.

- ✓ R_g stands for the resistance of the galvanometer.
- ✓ R_X represents a high resistance component called series resistance.
- ✓ G denotes the galvanometer coil.
- ✓ I is the total current flowing through the circuit.
- ✓ I_g signifies the total current passing through the galvanometer, corresponding to a full scale deflection.
- ✓ V represents the voltage drop across the series connection of the galvanometer and the high resistance.

  When the current I_g flows through the series combination of the galvanometer and the highresistance R, the voltage drop across the branch ab can be expressed as:

$$V = V_g + V_x$$

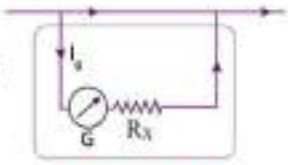


Figure 18. 15 Voltmeter and its schematic diagram

Current in series combination remain conserved therefore;

$$I_g = I_X$$

$$V = I_g R_g + I_g R_X$$

$R_x = V - I_g R_g / I_g$

$R_x = \frac{V}{I_g} - R_g$

This equation can be used to determine the value of R_X.

Worked Example 18.7

A galvanometer coil of 40Ω resistance shows full range deflection for a current of 4mA. How can this galvanometer be converted into a voltmeter of range 0-12V?

Solution:

Step 1: Write down the known quantities and quantities to be found.

$R_g = 40\Omega$,

V = range**0-12V**

$I_g = 4mA = 4\times10^{-3}A$

Step 2: Write down the formula and rearrange if necessary.

$V = I_g(R_g+R_x)$

Step 3: Put the values and calculate.

$$R_X = \frac{V}{I_g - R_g}$$

$R_X = (12/(4\times10^{-3}))-40$

$R_X = 2960\Omega$

Result: we get $\mathbf{R_X = 2960\Omega}$

SUMMARY

- Magnetic Field: The area around a permanent magnet or current-carrying conductor where a magnetic force is generated is called a magnetic field.
- Magnetic Force: The force exerted on a wire carrying current I in a uniform magnetic field is given by the formula: F=BILSinθ
- Measuring Magnetic Flux Density: A current balance can be used to measure magnetic flux density.
- Ampere's Law: This law quantifies how a current creates a magnetic field around it, providing a mathematical relationship between current and magnetic field strength.
- Solenoid: A long, tightly wound coil of wire used to generate a uniform magnetic field within the coil's interior. The magnetic field inside a solenoid is given by: $B=\mu_0 nI$
- Toroid: A solenoid curved to form a circular shape. The magnetic field strength inside a toroid is given by: $B=\frac{(\mu oNI)}{2\pi r}$. This formula shows that the magnetic field strength is inversely related to the radius rand directly related to the number of turns per unit length N and the current I passing through the turns.
- The shunt resistance for a ammeter is given by

$$R_s = \frac{R_g I_g}{I - I_g}$$

- In case of voltmeter the high resistance is given by

$$R_X = V - \frac{I_g R_g}{I_g}$$

4. The charge on a particle is doubled, and its velocity remains the same, how then the magnetic force on the particle will be:
   (a) doubled (b) halved
   (c) Is the same. (d) quadrupled
5. Strength of magnetic field of solenoid can be increased by:
   (a) Increasing number of turns
   (b) decreasing number of turns
   (c) increasing the current through the solenoid
   (d) inserting a ferromagnetic core (e.g., iron) into the solenoid
6. The magnetic field inside along solenoid is:
   (a) equal to zero
   (b) uniform
   (c) decreases as we go away from the center to surface
   (d) increases as we go to wards the surface
7. The force between two current carrying conductor arises due to
   (a) Magnetic effect of current (b) Polarization
   (c) Electromagnetic induction (d) Electrostatic interaction
8. To measure a higher voltage, what should you do with the voltmeter's internal resistance?
   (a) Increase it (b) Decrease it
   (c) Keep it the same (d) It doesn't affect the measurement
9. A proton moves perpendicular to a uniform magnetic field. What is the direction of the force experienced by the proton?
   (a) Parallel to the magnetic field (b) In the direction of the proton's velocity
   (c) Perpendicular to the magnetic field and the proton's velocity
   (d) Opposite to the direction of the proton's velocity
10. An electron moves parallel to a uniform magnetic field. What is the magnitude of the force experienced by the electron?
   (a) Maximum, since the electron is moving in the same direction as the field
   (b) Minimum, since the electron is moving perpendicular to the field
   (c) Zero, since the electron is moving parallel to the field
   (d) Depends on the speed of the electron

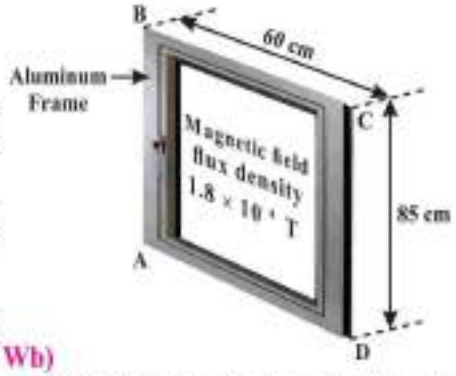
Section (B): CRQs (Short Answered Questions):

1. Charge particles are fired in vacuum tube hit a fluorescence screen. Will it be possible to know whether they positive or negative?
2. What is a solenoid, and how does it differ from a simple coil of wire?
3. Can a solenoid generate a magnetic field without any current flowing through it? Why or why not?
4. Explain why a toroid is often preferred over a straight solenoid when designing certain types of electrical components.
5. What role does the Ampere's circuital law play in understanding them agnetic field inside a toroid?
6. What is a galvanometer, and what is its primary function in an electrical circuit?

Section (C): ERQs (Long Answered Questions):

1. Can a galvanometer measure both DC and AC currents? Explain any limitations it might have with AC measurements.
2. Why should an ammeter ideally have a very low resistance compared to the circuit it is measuring?
3. What is the potential risk of using a voltmeter with a high internal resistance in a circuit? How can this risk be mitigated?
4. Describe the basic working principle of a voltmeter. How does it measure voltage across a component in a circuit?

Section (D): Numerical:



1. An aluminum window has a width of 60 cm and length of 85 cm as shown in the figure.
   a. When the window is closed the magnetic flux density is 1.8×10^{-4} Tisnormal to window.
   b. Calculate the magnetic flux through the window. **(9.18×10^{-5} Wb)**
2. The poles of a horse shoe magnet measures 8Cm × 3.2cm.the magnetic flux density between the magnet poles is 80mT. Outside of the magnet the magnetic flux density is zero. Calculate the magnetic flux density between the poles of a magnet.**(2.048×10^{-4} Wb)**
3. A wire 1.80 m long carries a current of 13.0 A and makes an angle of 35.0° with a uniform magnetic field of magnitude B = 1.50T. Calculate the magnetic force on the wire. **(19.46 N)**
4. A solenoid has length L = 1.23 m and inner diameter d = 3.55 cm, and it carries a current I= 5.57 A. It consists of five close-packed layers, each with 850 turns along length L. What is B at its center? **(Ans:24.2mT)**
5. A moving coil galvanometer has resistance of 50 Ω and it gives full scale deflection at4mA current. A voltmeter is made using this galvanometer and a 5KΩ resistance. Calculate the maximum voltage that can be measured using this voltmeter. **(Ans:20.2V)**
6. Compute the magnitude of the magnetic field of a long, straight wire carrying a current of 1A at distance of 1m from it. Compare it with Earth's magnetic field.
   (Ans:$2\times10^{-7}$and 100times than earth field)
7. Find the current in a long straight wire that would produce a magnetic field twice the strength of the Earth's at a distance of 5.0cm from the wire. (Magnetic field of Earth $=5.0\times10^{-5}$T). **(Ans:25A).**
8. What is flux density atadistanceof0.1m in air from along straight conductor carrying a current of 6.5 A. calculate the force per on a similar parallel conductor at a distance of 0.1m from the first and carrying a current of 3A. **(Ans: 13×10^{-6}Weber/m^{2}39$\times10^{-6}$N).**

Introduction:

Electromagnetic induction stands as a cornerstone in the realm of physics, serving as a fundamental principle that explains the dynamic relationship between electricity and magnetism. Discovered by the famous physicist Michael Faraday in the early 19th century, electromagnetic induction encompasses the phenomenon where the change in the magnetic fields induces the generation of an electromotive force (EMF) or voltage within a conductor. This revolutionary concept not only laid the groundwork for the understanding of essential electrical processes but also paved the way for the development of modern technologies such as electric generators and transformers. As we examine the intricacies of electromagnetic induction, we work out the interplay between magnetic flux, electric current, and the connection that binds these phenomena together. In this chapter, we understand the implications and applications of electromagnetic induction, from its beginning to its current prevailing influence on modern technology and industry.

19.1.2 Production of electricity by magnetism:

In order to demonstrate that moving electric charges (i-e an electric current) give rise to magnetic force, put a magnet compass near to current carrying wire the needle of compass will be deflected as shown in figure 19.1 This shows that moving charges produce magnetic field and in turn deflect magnetic compass. Such a demonstration for the first time was carried out in 1819 by Professor Hans Oersted of the University of Copenhagen

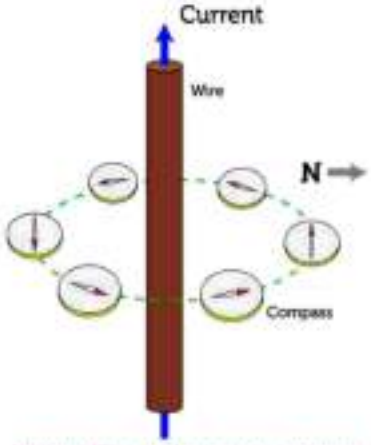


Figure 19.1 compass near to current carrying wire

The working of electric motor is based on this simple phenomenon, the current carrying coil produce magnetic field which interacts with the surrounding magnets causing the coil to rotate. In this case the electric energy is being converted into mechanical energy. The mechanical energy can be used to do variety of work, for example to pump water from ground level to a certain height.

On the other hand, the moving magnet produces electric force. To observe this phenomenon connect a coil with a sensitive galvanometer and a magnet move inside it. The relative motion of coil and motion will produce induced electric current. If the motion of magnet is stopped induced current ceases to exist.

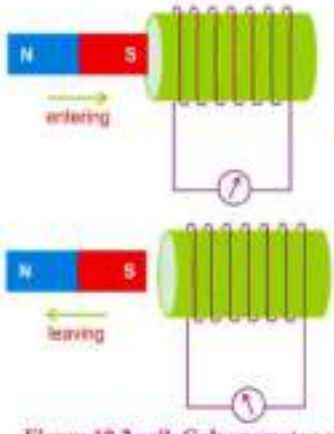


Figure 19.2 coil, Galvanometer and moving magnet

The simple experiment as shown in figure 19.2 is similar to the working of electric generator. In the electric generator the conductor coil is rotated within the field of permanent magnet. The changing magnetic field produces induced electric current. In 1831 Micheal Faraday

demonstrated that copper coils exposed to changing magnetic field produces electric current in the coil. Michael Faraday demonstrated this phenomenon with the help of copper coil, a magnet and a galvanometer. If the magnet was brought near to the coil an induced emf was produced and galvanometer shown current in one way. When the magnet was moved from the coil away the direction of induced current was reversed and galvanometer needle swing other way.

The phenomenon in which an emf is induced in the coil due to the change in linking magnetic flux is called electromagnetic induction.

DO YOU KNOW?

An induction stove, also known as an induction cooktop, is a type of cooking appliance that uses electromagnetic induction to heat cookware directly. Not all cookware is compatible with induction stoves. Cookware must have a magnetic bottom or base to work effectively on an induction cooktop.

19.1.1 Induced Electromotive force EMF:

Induced electromotive force (emf) can be generated in two primary ways: by relative movement and by changing a magnetic field.

(i) By Relative Movement (The Generator Effect):

This method is based on Faraday's Law of Electromagnetic Induction, which states that an emf is induced in a conductor when it experiences a change in magnetic flux. The generator effect, also known as motional emf, occurs when there is a relative motion between a conductor and a magnetic field.

Example:

In a simple electrical generator, a coil of wire rotates within a magnetic field. This induced emf drives an electric current through an external circuit.

(ii) By Changing a Magnetic Field (The Transformer Effect):

This method also relies on Faraday's Law but involves changing the strength or orientation of a magnetic field around a stationary conductor.

Example:

In a transformer, an alternating current in the primary coil creates a changing magnetic field. This is the basis for how transformers transfer electrical energy between circuits at different voltages.

Generator Effect:

Induced emf due to relative motion between a conductor and a magnetic field.

Transformer Effect:

Induced emf due to a changing magnetic field around a stationary conductor.

Both effects are fundamental principles in electromagnetism and are widely utilized in various electrical devices and systems.

19.1.4 Faraday's law of electromagnetic induction:

Faraday's law of electromagnetic induction is a fundamental principle in electromagnetism that describes the relationship between a changing magnetic field and the induced electromotive force (emf) or voltage in a closed circuit as illustrated in figure 19.3. It can be stated as follows:

The electromotive force (emf) induced in a closed circuit is directly proportional to the rate of change of magnetic flux passing through the circuit.

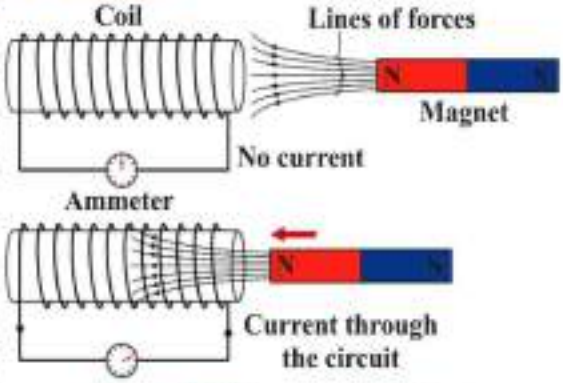


Figure 19.3 Electromagnetic induction

Mathematically, Faraday's law is often expressed as:

$$\varepsilon = -N\frac{\Delta\phi}{\Delta t} \quad .....(19.1)$$

Where ε represents the induced electromotive force (emf) measured in volts (V).

$\frac{\Delta\phi}{\Delta t}$ Represents the rate of change of magnetic flux with respect to time, it is measured in weber per second (Wb/s or V), and N indicates number of turns of the coil: Negative sign is introduced because the induced emf opposes the change in flux. It will be explained in Lenz's law.

DO YOU KNOW?

Electromagnetic braking is a method of slowing down or stopping a moving object, such as a vehicle, using the principles of electromagnetic induction. Electromagnetic braking is used in electric and hybrid vehicles, where regenerative braking systems utilize this technology to transform kinetic energy into electrical energy. This generated electrical energy is then stored in batteries or capacitors for future use.

19.1.3 Factors affecting the magnitude of the induced emf:

The magnitude of the induced electromotive force (emf) in a circuit or coil is governed by several factors, which are primarily described by Faraday's law of electromagnetic induction. These factors include:

1. **Magnetic Flux Change:** If the rate of change of magnetic flux is higher, it will lead to a larger induced emf. Conversely, a slow change in the magnetic field will result in a smaller induced emf.
2. **Number of Turns in the Coil:** If a coil of wire has more turns, the greater will be the induced emf. Each turn of the coil contributes to the emf, so increasing the number of turns increases the overall emf.
3. **Area of the Coil:** The size of the coil or the area it encloses also affects the induced emf. A larger coil or a coil with a larger cross-sectional area will capture more magnetic flux lines, resulting in a larger induced emf.

4. **Angle between Magnetic Field and Coil:** The angle between the magnetic field lines and the plane of the coil affects the induced emf. When the magnetic field is perpendicular to the coil's plane (90°), the emf is maximized. When the field lines are parallel (0°), the emf is minimized.
5. **External Factors:** External factors, such as temperature, material property, pressure, and other environmental conditions, can also influence the induced emf, especially in situations where materials may exhibit nonlinear magnetic properties.

Self-Assessment Questions:

1. Explain Faraday's law of electromagnetic induction and its significance in electrical systems.
2. How does the rate of change of magnetic flux affect the induced electromotive force (EMF) in a coil?

19.1.4 Lenz's law and principle of conservation of energy:

After the introduction of Faraday's law of electromagnetic induction, Heinrich Friedrich Lenz formulated a rule for determining the direction of an induced current within a loop.

According to Lenz's law ***"The direction of induced emf in a circuit is always such that it opposes the cause which produces it"***

In Figure 19.4 (a) as the North Pole of the magnet approaches the surface of the coil, an induced current is generated within the coil in a anti clock wise direction. Consequently, the coil generates a North magnetic pole, resulting in a repulsive force experienced by the approaching magnet. In this manner, the coil resists the factor responsible for inducing the electromotive force (emf).

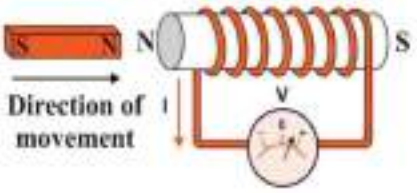


Figure 19.4 (a) Direction of induced emf

In Figure 19.3 (b), as the North Pole of the magnet moves away from the coil, an induced current is generated within the coil in a clockwise direction. Consequently, the coil generates a South magnetic pole, resulting in a attractive force experienced by the leaving magnet. Hence the law states that the induced current flows in a manner that oppose the change which is giving rise to it. That is why negative sign is introduced in equation 19.1.

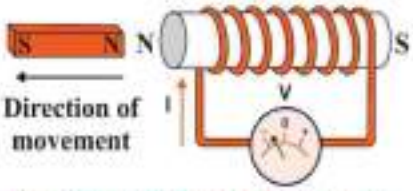


Figure 19.4 (b) Direction of induced emf

Lenz's law and Conservation of Energy:

The principle of conservation of energy states that energy cannot be created or destroyed; it can only change forms. When we apply this principle to electromagnetic induction, we see that the work done in changing the magnetic field is converted into electrical

energy in the form of the induced current. Hence lenz's law is in accordance with the law of conservation of energy.

Self-Assessment Questions:

1. Provide an example from everyday life or technology where Lenz's law is applied.
2. Discuss the consequences of not following Lenz's law in electromagnetic systems.

Worked Example 19.1

A coil with 100 turns and an area of 0.02 square meters is placed in a changing magnetic field. The magnetic field changes from 0.2 Tesla to 0.5 Tesla in 0.1 seconds. Calculate the induced electromotive force (emf) in the coil, the direction of the induced current, and how this relates to Lenz's law.

Solution:

Step 1: Write the known quantities and point out quantities to be found

B changes from 0.2 Tesla to 0.5 Tesla
Time t = 0.1 seconds.
No of turns = 100
Area A = 0.02 m^2
Induce emf = ?

Step 2: Write the formula and rearrange if necessary

$$\varepsilon = -N\frac{\Delta\emptyset}{\Delta t}$$

Step 3: Now, substitute the values into the formula:

$$\varepsilon = -N\frac{\emptyset_2 - \emptyset_1}{\Delta t} = 100 \times \frac{0.01 - 0.004}{0.1}$$

$$\varepsilon = -6\ V$$

Result: The induced emf in the coil is *6 V*, and the induced current flows counterclockwise to oppose the increase in the external magnetic field, in accordance with Lenz's law.

Worked Example 19.1

A coil with 200 turns and an area of 0.03 square meters is placed in a uniform magnetic field of 0.4 Tesla. The coil is quickly pulled out of the magnetic field in 0.2 seconds. Calculate the induced electromotive force (emf) in the coil and determine the direction of the induced current.

Solution:

Step 1: Write the known quantities and point out quantities to be found

B = 0.4T
Radius of coil = r =0.1m
$A = 0.03\ m^2$
The coil is pulled out of the magnetic field in 0.2 seconds.
Induce and its direction =?

Step 2: Write the formula and rearrange if necessary

Calculate the initial and final flux:

So initial flux = $\emptyset_1 = 0.4 \times 0.03 = 0.012$ Weber

Calculate the final flux:

When the coil is completely out of the magnetic field, the flux becomes zero because there is no magnetic field passing through the coil. Final flux $\emptyset_2 = 0$

Step 3: Now, calculate the rate of change of flux: =

$$\frac{\Delta\emptyset}{\Delta t} = \frac{\emptyset_2 - \emptyset_1}{\Delta t} = \frac{0 - 0.012}{0.2} = -0.06\, Wb/s$$

$$\varepsilon = -N\frac{\Delta\emptyset}{\Delta t} = -1.2\, Volts$$

Result: The induced current flows counterclockwise to oppose the decrease in magnetic flux, in accordance with Lenz's law.

Eddy currents and their magnetic and heating effects:

Eddy currents are circulating currents induced in a conductor when it is exposed to a changing magnetic field. They are a common phenomenon in electromagnetic systems. These currents can have both magnetic and heating effects. Now let's discuss about the production of eddy currents and their effects:

19.2.1 Production of Eddy Currents:

Eddy currents are produced according to Faraday's law of electromagnetic induction, which state that a changing magnetic field induces an electromotive force (emf) in a conductor. When a conductor, such as a metal plate or a coil of wire, is subjected to a changing magnetic field, the magnetic flux through the conductor changes, leading to the induction of eddy currents. As shown in figure 19.5.

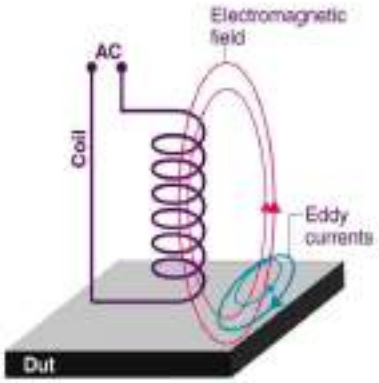


Figure 19.5
Production of Eddy Currents.

19.2.2 Magnetic Effects of Eddy Currents:

Counteracting Magnetic Field:

Eddy currents generate their own magnetic fields, and the direction of these fields opposes the original magnetic field that induced them as a result, eddy currents create a magnetic field that counteracts the original magnetic field's change, thereby reducing the net magnetic field.

Magnetic Damping:

In applications like electromagnetic brakes and magnetic dampers, eddy currents are intentionally induced to create a magnetic resistance that opposes motion. This magnetic damping effect is useful for controlling the movement of objects and slowing them down.

19.2.3 Heating Effects of Eddy Currents:

Joule Heating:

Eddy currents experience resistance as they flow through the conductor, and this resistance results in the conversion of electrical energy into heat, following Joule's law. Eddy currents can heat up the conductor. In some cases, such as induction heating for metal processing or cookware, this heating effect is used for practical applications.

19.2.4 Reduction of Eddy Currents:

In many electrical systems, eddy currents represent an undesirable source of energy loss, especially in transformers and electric motors. To minimize these losses, laminated cores are used in transformers to break up the conducting paths and reduce the formation of eddy currents.

The use of laminated iron cores in electric motors, generators, and transformers is essential for several important reasons:

In all three types of devices, the iron cores are exposed to alternating magnetic fields. When a solid iron core is used, it can create circular electrical currents within the core material called eddy currents. These eddy currents dissipate energy in the form of heat and can be highly inefficient. Laminated cores are constructed by stacking thin sheets or laminations of iron separated by insulating material. This design significantly reduces the formation of eddy currents because the laminations are electrically insulated from each other. Consequently, energy losses due to eddy currents are minimized.

Enhanced Efficiency:

By reducing the losses associated with eddy currents, laminated cores improve the overall efficiency of electric motors, generators, and transformers. Less energy is wasted as heat, allowing these devices to operate more efficiently and with less energy consumption.

Mitigation of Vibration and Noise:

Eddy currents generated in a solid iron core can lead to vibrations and noise, which can be undesirable in many applications. Laminated cores help reduce these vibrations and noise levels, making the devices quieter and more mechanically stable.

Better Cooling and Thermal Management:

Since laminated cores result in reduced heat generation due to decreased eddy current losses, they often allow for more efficient cooling and thermal management in these devices. This can lead to longer operational lifetimes and improved reliability.

Self-Assessment Questions:

1. When might the heating effects of eddy currents be undesirable?
2. How do the dimensions of a conductor impact the formation and magnitude of eddy currents?

19.3.1 Self Induction:

When an electric current pass through a coil, it creates a magnetic field around it. If the current in the coil changes, either increasing or decreasing, the magnetic flux within the coil also changes accordingly. As a result of the change in magnetic flux, an induced emf is generated in the same coil. This process, where a changing current in the coil induces an emf in the coil itself, is known as self-induction.

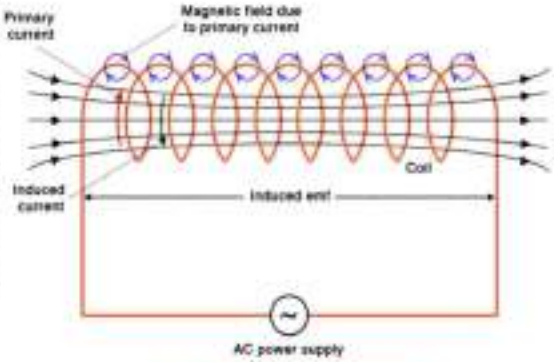


Figure 19.6 Self Inductions

Consider the circuit shown in Figure 19.6. which comprises a coil connected in series with a battery and a rheostat. When the rheostat is adjusted, it changes the current flowing through the circuit, leading to changes in the magnetic flux within the coil. These variations in magnetic flux result in the induction of an electromotive force (emf) within the coil.

If we denote the magnetic flux through a single loop of the coil as $\emptyset$, then the total magnetic flux passing through the N turns of the coil would be $N\emptyset$. This relationship holds true because the magnetic flux, directly proportional to the magnetic field, which, in turn, is proportionate to the current (I) flowing through the circuit

$$N\emptyset \propto I$$

$$N\emptyset = LI \ldots\ldots\ldots\ldots\ldots (19.2)$$

Where $L = \frac{N\emptyset}{T}$ is constant of proportionality called the self inductance of coil. It depends on factors such as the coil's number of turns, its cross-sectional area, and the material used for the core. Hence inductance of coil made up of soft iron is greater than the air core coil.

By the Faradays law of electromagnetic induction

$$\varepsilon = -N\frac{\Delta\emptyset}{\Delta t} \ldots\ldots\ldots\ldots\ldots\ldots (19.3)$$

Substitute equation (19.2) in Equation (19.3), we get

$$\varepsilon = -L\frac{\Delta I}{\Delta t} \ldots\ldots\ldots\ldots\ldots (19.4)$$

Equation (19.4) shows that the self-induced electromotive force (emf) within a coil is directly proportional to the rate of change of current within the coil. We can define the self-inductance, denoted as "L," as the ratio of this emf to the rate of change of current in the coil. The negative sign in equation (19.4) signifies that the self-induced emf always opposes the cause that produces it. This is the reason why the self-induced emf is often referred to as the "back emf." Consequently, due to the phenomenon of self-inductance, coils of wire are commonly known as "inductors." In the field of electrical and electronics engineering, inductors find widespread use.

Unit of Inductance:

The unit of self-inductance is the Henry (H), named after the American physicist Joseph Henry.

One Henry (1 H) is defined as the amount of self-inductance in a circuit when a change in current of 1 ampere per second (1 A/s) induces an electromotive force (emf) of 1 volt (1 V) within the same circuit. Mathematically, this relationship can be expressed as:

$$1\ Henry = \frac{1 Volt}{1\ Ampere\ per\ Second}$$

19.3.2 Energy Stored in an Inductor:

An inductor is a passive electrical component that stores energy in the form of a magnetic field when an electric current flow through it. The ability of an inductor to store electric potential energy is based on the fundamental principle of electromagnetic induction. Here's a simple explanation of how an inductor stores electric potential energy:

Current Flow:

When an electric current flow through a coil of wire (the inductor), it creates a magnetic field around the coil.

Magnetic Field Buildup:

As the current increases, the strength of the magnetic field around the coil also increases. This process takes a short amount of time, as the magnetic field does not build up instantly; it grows with the rate of change of the current.

Energy Storage:

The energy is stored in the magnetic field. The inductor stores electric potential energy in this magnetic field. The more current that flows through the inductor or the faster the current changes, the stronger the magnetic field, and thus, the more energy is stored.

Magnetic Field Collapse:

When the current decreases or stops (like when you turn off the power), the magnetic field starts to collapse. As it collapses, it induces an electromotive force (EMF) or voltage in the coil.

Released Energy:

This induced voltage represents the stored energy being released. The inductor converts the stored magnetic energy back into electric energy. This energy can be used to sustain the current for a short period or can be transferred to other parts of the circuit.

Self-Assessment Questions:

1. When is the energy stored in an inductor released?
2. Provide examples of practical applications where the energy stored in inductors is crucial.

19.3.3 Energy produced in Self Induction:

An inductor stores energy in its magnetic field when it carries a direct current (DC). This energy remains stored as long as the inductor continues to carry the current. When the current in the inductor increases, the stored energy also increases, and conversely, it decreases when the current is reduced.

Consider an inductor connected to a DC power source through a switch, as depicted in Figure 19.7. When the switch is closed, the current in the inductor gradually rises until it reaches its maximum value, denoted as I. This changing current leads to a corresponding change in the magnetic flux within the coil, causing an induced electromotive force (emf) to establish itself in the coil. Consequently, an induced current is generated in the circuit, which works to minimize the current produced by the battery in accordance with Lenz's law.

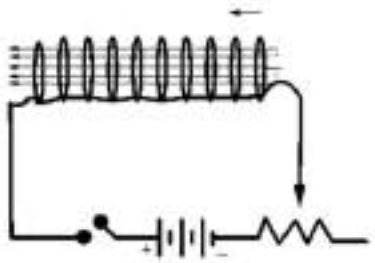


Figure 19.7 Energy produced in Self Induction

As a result, the battery must perform work on the charges to build up the current I. This work is mathematically expressed as follows:

$$W = \Delta V \Delta q \quad \ldots\ldots\ldots\ldots (19.5)$$

As the induced emf is produced in the inductor so $\Delta V = \mathcal{E}_L = L\frac{\Delta I}{\Delta t}$ and using work energy relation equation (19.5) can be written as

$$E = \frac{\Delta q}{\Delta t} L \Delta I \quad \ldots\ldots\ldots\ldots (19.6)$$

Since $\frac{\Delta q}{\Delta t} = \frac{0+I}{2} = \frac{I}{2}$

Equation (19.6) becomes $E = \frac{1}{2}LI^2$

Where:

- E is the energy stored in joules (J).
- L is the self-inductance of the inductor in Henries (H).
- I is the current flowing through the inductor in amperes (A).

This formula tells you that the energy stored in an inductor is directly proportional to the square of the current passing through it and is also dependent on the inductance of the inductor itself. As the current increases, the energy stored in the inductor increases, and as the current decreases, the stored energy decreases.

19.4.1 Mutual Inductance:

Imagine two closely positioned coils, as depicted in Figure 19.8. The first coil, known as the primary coil, is connected to a battery through a variable resistor 'R' and a switch 'S.' The second coil, known as the secondary coil, is connected to a galvanometer.

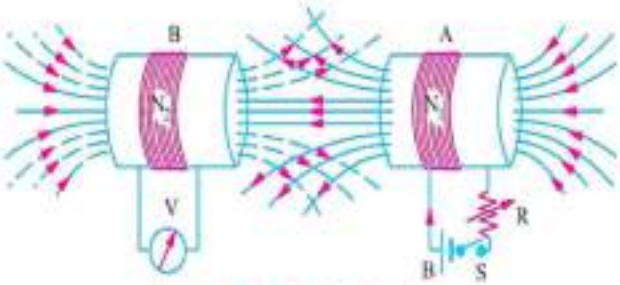


Figure 19.8 Mutual Inductance

When the resistance is adjusted, it causes a change in the

electric current 'I' within the primary coil. This change results in the production of a varying magnetic field that interacts with the nearby secondary coil, effectively establishing a connection between them. Consequently, an electromotive force (EMF) is induced in the secondary coil in accordance with Faraday's law. **"A phenomenon where a changing current in one coil induces an electromotive force (EMF) in another coil is known as mutual induction".**

Consider the two closely wound coils of wire shown in Figure 19.8. The current I_1 in primary coil A, which has N_1 turns, creates a magnetic field. Some of the magnetic field lines pass through secondary coil B, which has N_2 turns. Let the flux passing through secondary coil having N turns be is $N_s\emptyset_s$. This magnetic flux depends on the primary coil current I_p

$$N_s\emptyset_s \propto I_p$$

$$N_s\emptyset_s = MI_p \ldots\ldots\ldots\ldots (19.7)$$

$$M = \frac{N_s\emptyset_s}{I_p}$$

Here, 'M' represents the proportionality constant known as the mutual inductance between the two coils. Mutual inductance depends on factors such as the number of turns in the coils, their geometrical characteristics, the materials used to construct the coils, and the separation between them. When the separation between the coils increases, the mutual inductance decreases as a result of the diminishing magnetic flux.

From faradays law of electromagnetic induction

$$\varepsilon_s = -N_s\frac{\Delta\emptyset_s}{\Delta t} \ldots\ldots\ldots (19.8)$$

Substitute Equation (19.7) in Equation (19.8), we get

$$\varepsilon_s = -M\frac{I_p}{\Delta t} \ldots\ldots\ldots\ldots\ldots (19.9)$$

Equation (19.9) shows that induced emf produce in the secondary coil is equal to negative rate of change of current in the primary coil.

Equation (19.9) can be used to provide us a definition of mutual inductance which is Henry

$$M = -\frac{\varepsilon_s}{\left(\frac{I_p}{\Delta t}\right)}$$

One Henry (H) of mutual inductance can be defined as the mutual inductance between two coils when an EMF of one volt is induced in the secondary coil when the rate of change of current in the primary coil is one ampere per second.

Mutual inductance measures the ability of one coil to induce an electromotive force (EMF) in another coil when the current in the first coil changes.

Self-Assessment Questions:

1.Give an example of devices or systems where mutual induction is utilized.

2. What are the advantages of using mutual induction in practical applications.

19.5.1 Transformer:

A transformer is an electrical device used to transfer electrical energy between two or more coils of wire through electromagnetic induction. It can either step up (increase) or step down (decrease) the voltage of an alternating current while keeping the frequency of the alternating current unchanged. It works on the principle of mutual induction.

Construction of a Transformer:

The transformer consists of following main components as shown in figure 19.9.

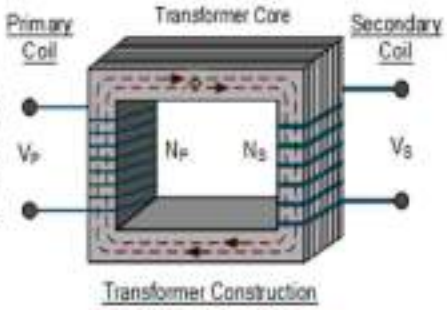


Figure 19.9 Transformer:

Core:

Transformers consist of two coils of wire, known as the primary coil and the secondary coil, wound around a common magnetic core. The core is typically made of materials with high magnetic permeability, such as laminated iron or ferrite, to enhance the efficiency of the transformer.

Primary Coil:

The primary coil is connected to the input voltage V_p source. It consists of a specific number of turns of wire wound around one section of the core denoted as N_p.

Secondary Coil:

The secondary coil is connected to the load or the device that needs the transformed voltage. It has a different number of turns of wire wound around another section of the core denoted by N_s.

Insulation:

Both the primary and secondary coils are insulated from each other to prevent electrical contact and ensure electrical isolation.

Working of a Transformer:

Imagine we have an alternating electromotive force (emf) applied to the primary coil. If, at a certain moment (t), the magnetic flux within the primary coil is changing at a rate of $\frac{\Delta\phi}{\Delta t}$ this change in flux will induce a counteracting electromotive force (emf) in the primary coil, opposing the applied voltage. We can express the instantaneous value of this self-induced emf as follows:

$$\text{Self-induced emf} = -N_p \frac{\Delta\phi}{\Delta t}$$

If the coil's resistance can be considered negligible, then the opposing emf in the primary coil is equal in magnitude but opposite in direction to the applied voltage. This relationship can be represented as:

$$V_p = -N_p \frac{\Delta\phi}{\Delta t} \quad \text{.....} \quad (19.10)$$

Here, N_p represents the number of turns in the primary coil.

Now, let's assume that the flux passing through the primary coil also flows through the secondary coil. In other words, these two coils are magnetically linked. Hence, the rate of change of magnetic flux in the secondary coil will also be $\frac{\Delta\emptyset}{\Delta t}$, and the magnitude of the induced emf across the secondary coil can be expressed as:

$$V_s = -N_s \frac{\Delta\emptyset}{\Delta t} \ldots\ldots\ldots\ldots\ldots (19.11)$$

Where *Ns* is number of turns in secondary

When we divide equation (19.11) by equation (19.10), we obtain

$$\frac{V_s}{V_p} = \frac{N_s}{N_p}$$

Step-Up or Step-Down Transformers:

The ratio of the number of turns in the primary coil (Np) to the number of turns in the secondary coil (N_s) determines whether the transformer is a step-up or step-down transformer. If $N_s > N_p$ than it's a step-up transformer and it increases the voltage. Conversely, if $N_p > N_s$ it is a step-down transformer and it decreases the voltage.

19.5.2 Step up transformer:

Step-up transformers are designed to increase the voltage of electricity. They have more turns in the secondary coil than in the primary coil, resulting in a higher secondary voltage compared to the primary voltage.

Reduced Current:

Increasing the voltage through a step-up transformer reduces the current while maintaining the same power (P = VI). Lower current reduces the resistive losses in the transmission cables. The power lost as heat (I^2R losses) is proportional to the square of the current, so decreasing the current significantly reduces these losses.

Efficient Long-Distance Transmission:

High voltage is essential for efficient long-distance power transmission. Step-up transformers are used at power generation plants to raise the voltage, allowing electricity to be transported over extensive electrical grids with minimal energy loss. This is especially important for transmission lines covering large distances.

19.5.3 Step-Down Transformers:

Voltage Reduction:

Step-down transformers, on the other hand, lower the voltage from high levels to safer, more usable levels for homes and industries. They have fewer turns in the secondary coil, which decreases the secondary voltage.

Balanced Current:

While step-down transformers decrease voltage, they increase the current to maintain the same power. This is beneficial for local distribution because it ensures that the power can be delivered to homes, businesses, and industrial facilities while minimizing I^2R losses over shorter distances.

3. Transmission and Substations:

The high-voltage electricity is transmitted over transmission lines to substations located at various points in the electricity distribution network. At these substations, step-down transformers are used to lower the voltage to a more manageable level for further distribution.

4. Step-Down Transformers in Substations:

Step-down transformers reduce the high-voltage electricity to lower, safer voltage levels, suitable for local distribution. This lower voltage is then distributed to houses through distribution lines.

Worked Example 19.3

Suppose a step-up transformer have 200 turns in the primary coil and 800 turns in the secondary coil. If the voltage in the primary coils is 120 volts, calculate the voltage in the secondary coil.

Solution:

Step 1: Write down the known quantities and quantities to be found.

$N_p = 200$ $N_s = 800$ $Vp = 120$ volts $Vs = ?$

Step 2: Write down the formula and rearrange if necessary.

Using the turns ratio equation: $\frac{V_s}{V_p} = \frac{N_s}{N_p}$

Step 3: Substitute the known values:

Substitute the known values: $\frac{V_s}{120} = \frac{800}{200}$

Step 4: Solve for V_s,

We get: $V_s = 480\ Volts$

Result: The voltage in the secondary coil is 480 volts.

Worked Example 19.4

Let a step-down transformer have 600 turns in the primary coil and 150 turns in the secondary coil. The voltage in the primary coil is 240 volts. Calculate the voltage in the secondary coil.

Solution:

Step 1: Write down the known quantities and quantities to be found.

$N_p = 600$ $N_s = 150$ $Vp = 240$ $Vs = ?$

Step 2: Write down the formula and rearrange if necessary.

$\frac{V_s}{V_p} = \frac{N_s}{N_p}$

Step 3: Substitute the known values:

$$\frac{V_s}{240} = \frac{150}{600}$$

Step 4: Now, solve for V_s, we get:

$V_s = 60$ Volts

Result: So, the voltage in the secondary coil is 60 volts.

19.6.1 Motional emf:

Motional electromotive force (emf) is a phenomenon that arises when a conductor moves through a magnetic field, inducing an electromotive force within the conductor. This concept is based on Faraday's law of electromagnetic induction and is a fundamental aspect of electromagnetism. Motional emf is commonly appears in various practical applications. For example, the generation of electricity by a generator. When a coil of wire rotates in a magnetic field, motional emf is induced, leading to the generation of electric current.

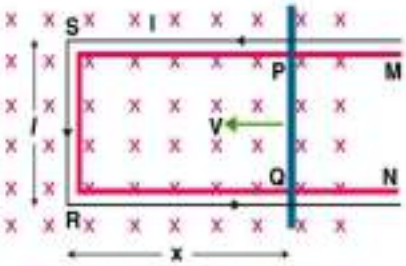


Figure 19.11 Motional emf

Understanding motional emf is crucial in designing and analyzing systems involving the motion of conductors in magnetic fields, and it plays an important role in the working of devices such as generators and certain types of sensors.

In summary, the Motional electromotive force (emf) is the voltage induced in a conductive rod as it moves through a magnetic field.

Consider a straight wire PQ having length l, as shown in figure 19.11. The conductor is in motion within a rectangular loop PQRS. This motion occurs within a uniform magnetic field B, perpendicular to the plane of the page. We assume that the rod moves uniformly at a constant velocity of v meters per second, and the surface it moves on is frictionless.

Each free electron of the conductor is moving with the conductor and thus experiences a force

$$F = -e\,(\vec{V}\times\vec{B}) \quad \text{or} \quad F = e\,(\vec{V}\times\vec{B})$$

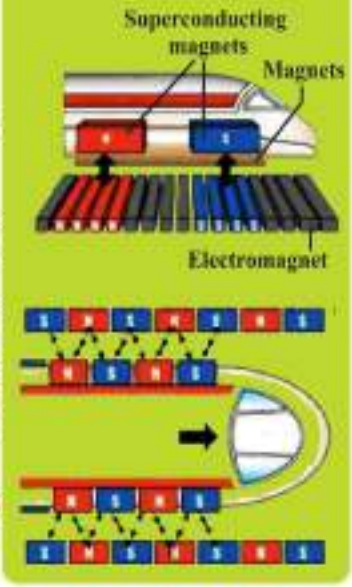
Electrons continue to accumulate at one end of the conductor, creating an excess of electrons, which results in a negative charge at that end. Conversely, the opposite end, with a deficiency of electrons, leads to the development of a positive charge. This process generates a potential difference within the wire, which is termed as motional electromotive force (emf). In order to derive an expression for the motional emf, we use the definition of potential difference.

Since the potential difference is defined as work done per unit charge

$$V = \frac{w}{q} = \frac{Fl}{q} = \frac{qvB\sin\theta \times l}{q} = vBl\sin\theta$$

DO YOU KNOW?

Magnetic Levitation (Maglev) fastest trains in the world use the repelling force of magnets to stay in the air. So, they do not even have wheels and also do not touch the track. Maglev trains use superconducting magnets and motional emf to levitate and propel the train along the tracks.



If the conductor is moving right angle to the magnetic field then $\theta = 90^0$

$$V = vBl \ldots\ldots\ldots\ldots 19.12$$

Where **V** is motional emf, v represents the velocity of conductor moving in a magnetic field and $\boldsymbol{l}$ is the length of conductor

Self-Assessment Questions:

1. Describe Fleming's right-hand rule and its application in the context of motional EMF.
2. How does the direction of motion, magnetic field, and current relate in Fleming's right-hand rule?

Worked Example 19.5

A straight conductor of length L = 0.5 meter move at a velocity of 2 m/s perpendicular to a magnetic field B = 0.1 Tesla. Calculate the motional emf induced in the conductor.

Solution:

Step 1: Write down the known quantities and quantities to be found.

Length of the conductor, $L = 0.5$ meter

Velocity of the conductor, $v = 2$ m/s

Magnetic field strength, $B = 0.1$ Tesla

Step 2: Write down the formula and rearrange if necessary.

$V = vBl$

Step 3: Substitute the known values:

$V = (0.1\ \text{T}) \cdot (0.5\ \text{m}) \cdot (2\ \text{m/s})$

$V = 0.1\text{V}$

Step 4: Therefore, the motional emf induced in the conductor is 0.1 volts.

Result: This means that as the conductor moves through the magnetic field, an electromotive force of 0.10 volts is induced across its ends, creating the potential for an electric current if the circuit is closed.

19.6.2 AC Generator

An AC generator, also known as an alternator, is a device that converts mechanical energy into electrical energy in the form of alternating current (AC). It's based on the principle of electromagnetic induction, which was first described by Michael Faraday.

Construction of an AC Generator:

An AC generator typically consists of the following four components as shown in figure 19.12:

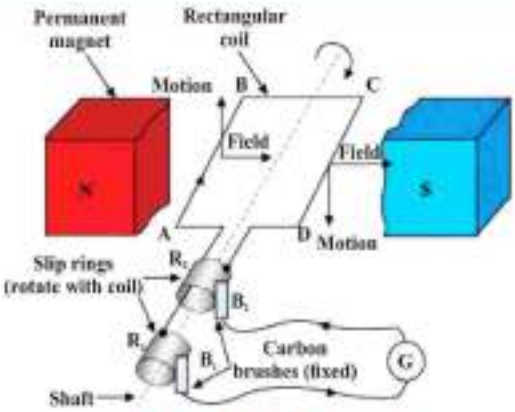


Figure 19.12 AC generator

(i) Armature (iii) Slip-rings
(ii) Field magnet (iv) Brushes

Armature:

The armature or rotor is a rectangular coil mounted on a rotating shaft. It is the component that spins within a magnetic field to induce electrical current.

Permanent magnet:

The permanent magnet or stator is a stationary part of the generator that produces a magnetic field. It is typically made up of a set of magnets or electromagnets arranged in a circular or cylindrical configuration around the rotor.

Slip Rings and Brushes:

The ends of the rotor coil are connected to slip rings, which are conductive rings that rotate with the rotor. Brushes, typically made of graphite, press against the slip rings to collect the generated electrical current.

Shaft and Bearings:

The rotor is mounted on a shaft that allows it to rotate freely. Bearings are used to reduce friction and enable smooth rotation.

Working of an AC Generator:

The working of an AC generator involves several steps:

Rotation of the Armature coil:

The coil or rotor is mechanically rotated using an external source of power, such as an engine, a turbine, or any other energy source.

Generation of a Changing Magnetic Field:

As the rotor spins within the stator's magnetic field, the magnetic field within the rotor coil changes. This changing magnetic field induces an electromotive force (emf) or voltage in the coil,

Generation of Alternating Current:

The induced voltage causes an electric current to flow through the coil, and since the magnetic field's direction is changing, the current produced is alternating in nature. This means the direction of the current constantly switches back and forth, resulting in an AC output.

Collection of Output:

The alternating current generated in the rotor coil is collected using slip rings and brushes. The brushes maintain contact with the slip rings as they rotate, allowing the generated AC to be drawn from the generator.

Induced EMF of an AC generator:

Suppose the armature coil AHCD rotates counterclockwise. As it rotates, the magnetic flux linked with it changes, inducing a current in the coil as shown in figure 19.13(a,b). The direction of the induced current follows Fleming's right-hand rule.

When the armature is in the vertical position and rotates counterclockwise, wire AH moves downward, while DC moves upward. This causes the induced electromotive force (emf) to flow from H to A and from D to C, within the coil, forming a current path of DCHA.

Stator:

The stator is the stationary part of the motor and contains the primary windings. Its main role is to produce a rotating magnetic field when AC voltage is applied. The rotating magnetic field interacts with the rotor to induce motion.

Rotor:

The rotor is the rotating part of the motor. It can be of different types, such as squirrel-cage or wound rotor, depending on the motor design. When the stator produces a rotating magnetic field, the rotor experiences a torque due to the interaction with the field. This torque causes the rotor to rotate and generate mechanical output.

Bearings:

Bearings are essential components that support and allow the rotor to rotate within the stator. They reduce friction, enabling smooth and efficient operation of the motor.

Shaft:

The shaft is connected to the rotor and extends beyond the motor housing. It transfers mechanical energy to the outside world, allowing the motor to perform useful work when connected to a load.

Cooling System:

Many AC motors incorporate cooling systems, such as fans or fins, to dissipate heat generated during operation. Efficient cooling is important to prevent overheating and prolong the motor's lifespan.

The main features and components work together to enable the AC motor to function. When AC voltage is applied to the stator windings, a rotating magnetic field is created, which exerts a torque on the rotor. The rotor's rotation results in mechanical work being performed by the motor, and the motor can be used to drive various mechanical devices, such as fans, pumps, conveyor belts, and more.

The efficiency and performance of an AC motor depend on the design and quality of its components, the load it drives, and its operating conditions. Proper maintenance and control mechanisms, such as speed control, are often used to optimize the motor's performance in different applications.

19.6.2 The production of back emf in electric motors:

The concept of back electromotive force (back EMF) in electric motors is fundamental to understanding motor operation and efficiency. Back EMF is a self-generated electromotive force that opposes the applied voltage in a motor. It plays a crucial role in motor behavior, especially in limiting current and regulating speed. Here's a detailed explanation of the concept of back EMF in electric motors:

Electromagnetic Induction:

Back EMF is a consequence of Faraday's law of electromagnetic induction, which states that a change in magnetic flux through a coil of wire induces an electromotive force (EMF) in that coil. In an electric motor, the coil of wire is typically part of the rotor or armature.

Rotating Magnetic Field:

When an electric motor is powered, the stator (the stationary part of the motor) generates a rotating magnetic field by applying an alternating current (AC) voltage to its windings. In a direct current (DC) motor, the commutator and brushes create a changing magnetic field.

Rotor or Armature Interaction:

The rotor (or armature) in the motor is mounted within the magnetic field created by the stator. As the rotor rotates, it cuts through the magnetic field lines, causing a change in magnetic flux within the coils of wire on the rotor.

Back EMF Generation:

The change in magnetic flux through the rotor windings induces a voltage, which is referred to as back EMF. The direction of the induced voltage is in opposition to the applied voltage. In other words, it generates a voltage that resists the flow of current through the motor windings.

Current Regulation:

Back EMF has a critical function in motor operation. As the motor speeds up, the back EMF increases. This increase in back EMF results in a decrease in the net voltage across the motor windings. Consequently, the current flowing through the motor decreases. This self-regulation mechanism is essential for preventing the motor from drawing excessive current, overheating, and potentially damaging itself.

Effect on Motor Speed:

The relationship between back EMF and motor speed is inverse. When the motor operates at a higher speed, the back EMF is greater, and the current is reduced, helping maintain a consistent speed. Conversely, at lower speeds or when subjected to a higher mechanical load, the back EMF decreases, allowing more current to flow and providing the necessary torque to overcome the load.

In summary, the concept of back EMF in electric motors is the self-induced voltage that opposes the applied voltage and helps regulate the current and speed of the motor. It is a critical factor in ensuring the safe and efficient operation of electric motors, and understanding it is essential for motor design, control, and performance analysis.

Self-Assessment Questions:

1. Discuss how the production of back EMF aligns with the principle of energy conservation.
2. What is the primary purpose of the back EMF in electric motors?

EXERCISE

Section (A): Multiple Choice Questions (MCQs)
Choose the correct answer:

1. Electricity production by magnetism involves
   (a) the conversion of chemical energy to electrical energy.
   (b) the use of static electricity for power generation.
   (c) the movement of conductors within magnetic fields.
   (d) the absorption of solar energy by photovoltaic cells.
2. Induced e.m.f can be generated by relative movement, known as the generator effect, and by changing magnetic fields, known as the transformer effect.
   (a) True in transformers only.
   (b) False in all cases.
   (c) Applicable to both cases.
   (d) True only for static magnetic fields.
3. The magnitude of induced e.m.f. increases with
   (a) the decrease in the speed of the conductor's motion.
   (b) the decrease in the strength of the magnetic field.
   (c) the increase in the number of turns in the coil.
   (d) the reduction in the area of the coil.
4. Faraday's law of electromagnetic induction states
   (a) that current in a circuit always opposes the change in magnetic flux.
   (b) that the induced e.m.f. is directly proportional to the rate of change of magnetic flux.
   (c) that a changing electric field produces a magnetic field.
   (d) that magnetic fields can only be produced by electric currents.
5. Lenz's law predicts the direction of an induced current
   (a) to conserve electric charge.
   (b) by stating that the induced current will oppose the change causing it.
   (c) by following the direction of the applied magnetic field.
   (d) based on the magnitude of the electric field.
6. Eddy currents produce:
   (a) static electricity in conductive materials.
   (b) alternating magnetic fields with no heating effects.
   (c) magnetic fields that oppose the inducing field and cause heating in conductors.
   (d) uniform electric fields that have no impact on power generation.
7. Laminated iron cores are used in electric motors and transformers to
   (a) reduce the weight of the device.
   (b) enhance the mechanical strength.
   (c) minimize eddy current losses and reduce heating.
   (d) increase the magnetic permeability of the core.

8. Self-induction is the phenomenon where
   (a) a changing electric field induces a magnetic field.
   (b) a changing magnetic field within a coil induces an emf in the same coil.
   (c) a constant magnetic field induces a constant emf.
   (d) two coils induce emf in each other through mutual induction.
9. An inductor stores electric potential energy in:
   (a) the electric field around it.
   (b) the magnetic field within its coil.
   (c) the capacitance of its windings.
   (d) the heat generated by its resistance.
10. Transformers work on the principle of
   (a) converting direct current (DC) to alternating current (AC).
   (b) electromagnetic induction between primary and secondary coils.
   (c) generating electricity through chemical reactions.
   (d) using permanent magnets to maintain a constant voltage.

Section (B): CRQs (Short Answered Questions):

1. Explain Faraday's law of electromagnetic induction.
2. State Lenz's law and discuss its significance in the context of electromagnetic induction.
3. How does a step-up transformer differ from a step-down transformer?
4. Describe the principle of operation of an AC generator.
5. Define self-induction and explain how it occurs in a coil.
6. Define mutual induction and provide an example of a system exhibiting mutual induction.
7. How does the arrangement of coils influence the degree of mutual induction between them?
8. Define motional electromotive force (emf)
9. Differentiate between motional emf and electromagnetic induction in terms of their fundamental principles.
10. Differentiate between AC (alternating current) and DC (direct current) in the context of long-distance power transmission.

Section (C): ERQs (Long Answered Questions):

1. How does the strength of the magnetic field affect the induced electromotive force (emf) in a coil according to Faraday's law? b. Explain the relationship between the magnetic field strength and the magnitude of the induced emf.
2. In Faraday's law, why is the area of the coil an important factor in determining the induced emf?
3. How does changing the orientation of a coil with respect to the magnetic field influence the induced emf? Also describe the role of the number of turns in a coil in the context of induced emf. How does increasing the number of turns in a coil affect the induced emf, and why?
4. According to Faraday's law, why is the rate of change of magnetic flux crucial for inducing emf? Discuss how the speed of motion or change in magnetic field influences the induced emf.

5. How does the resistance of the conductor impact the flow of induced current in a circuit? In what situations might the resistance of the conductor become a significant factor in induced emf?
6. Explain the relationship between the frequency of the changing magnetic field and the induced emf. How does the frequency of the AC source affect the generation of induced emf in a coil?
7. Discuss how the material properties of a coil, such as its conductivity, can influence induced emf. In what ways might the type of material used in a coil affect its response to changing magnetic fields?
8. Why is the presence of a core material significant in transformers concerning induced emf? Explain how the type and properties of the core material can impact the efficiency of a transformer.
9. Why are slip rings used in AC generators, and how do they differ from commutator used in DC generators.
10. Why a commutator is necessary in DC motors, and how does it facilitate the continuous rotation of the motor?
11. What are the common applications of DC motors in everyday devices?

Section (D): Numerical:

1. Two coils are placed adjacent to each other, and a change in current in the first coil induces an emf of 0.5 V in the second coil. If the mutual inductance is 0.2 H, calculate the rate of change of current in the first coil. **(2.5A/s)**
2. A coil with an inductance of 0.5 H experiences a rate of change of current of 2 A/s. Calculate the induced EMF in the coil. **(-1V)**
3. Two coils are placed close to each other. If a change in current of 3 A/s in the first coil induces an EMF of 4 V in the second coil, calculate the mutual inductance. **(1.33H)**
4. A transformer has 200 turns in the primary coil and 400 turns in the secondary coil. If the primary voltage is 120 V, calculate the secondary voltage for a step-up transformer. **(240V)**
5. An inductor with an inductance of 0.02H has a current flowing through it of 2 A. Calculate the energy stored in the inductor. **(0.04J)**
6. A coil stores energy in form of electric potential energy of 0.2J when it carries a current of 2A. Calculate the inductance of coil. **(0.1H)**
7. An AC generator produces an alternating current with a maximum voltage of 240 V. If the frequency of the generated AC is 50 Hz, calculate the peak value of the voltage. **(339V)**
8. A transformer has 1000 turns in its primary coil and 200 turns in its secondary coil. If the primary voltage is 120 V, calculate the secondary voltage for a step-down transformer. **(24V)**
9. A conductor of length 0.4 m moves at a velocity of 5 m/s perpendicular to a magnetic field of 0.3 T. Calculate the motional EMF induced in the conductor. **(0.6V)**

Introduction:

Alternating current is the back bone of electrical engineering and physics. In addition, we will explore the dynamic nature of AC current, which is quite different when comparing with its counterpart, Direct Current (DC).

There is great contribution of many scientists, like Nikola Tesla and George Westinghouse in better understanding of alternating current. Their groundbreaking work in the late 19th and early 20th centuries laid the foundation for the widespread use of AC power, revolutionizing the way electricity is generated, transmitted, and utilized.

In this chapter we will discuss essential concepts such as frequency, amplitude, and the sinusoidal waveform that characterizes AC current. We will use the mathematics behind AC circuits and study the principles governing the behavior of resistors, capacitors, and inductors when connected with alternating current source. Also, the practical applications of AC current, from the power distribution grid that lights up our cities to the metal detector that have become integral to our daily lives.

20.1 Alternating Current

A graphical representation, such as the one depicted in Figure 20.1, shows the nature of an alternating current. This graph visually demonstrates the cyclic variation of the current, depicting a pattern where, during one half of the cycle, the current is positive, while during the other half, it is negative. Consequently, the current alternates its direction in the wires it traverses, flowing first in one direction and then in the opposite direction.

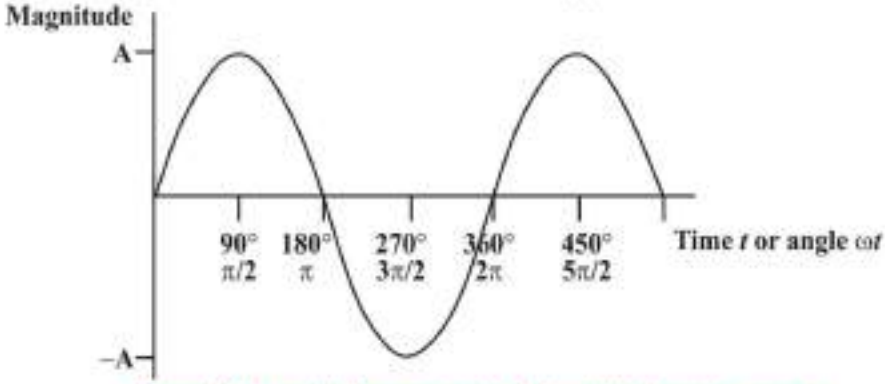


Figure 20.1 A graphical representation of sinusoidal alternating current.

In the context of utilizing household appliances connected to the mains, the current continuously oscillates back and forth along the wires connecting the user to the power station generating the electricity. At any given moment, the current possesses a specific magnitude and direction, as indicated by the graph.

Alternating current (AC) varies in a sinusoidal pattern and follows the shape of a sine wave. This pattern is characterized by a smooth, periodic oscillation that changes direction between positive and negative. While any current that alternates direction could be considered AC, the term typically refers to currents that exhibit this regular, sinusoidal pattern. This sinusoidal nature is fundamental to how AC is generated and utilized in power systems.

20.1.1 AC terminologies

Cycle:

One complete set of both positive and negative values of an alternating quantity is known as a cycle as shown in figure 20.2.

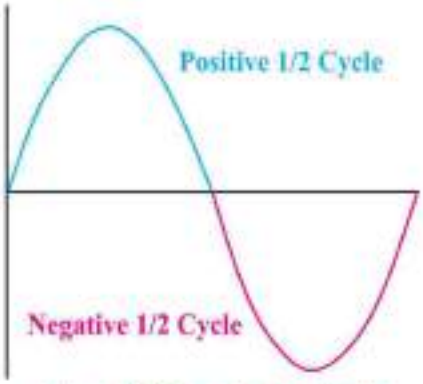


Figure 20.2 One cycle of ac current

Time Period (T):

The time period of an alternating current (AC) or voltage wave is the duration it takes to complete one full cycle. It is typically denoted by the symbol T. The relationship between time period and frequency (f) is given by the equation: $= \frac{1}{f}$, where T is in second.

Frequency (f):

It refers to the number of complete cycles of the current waveform that occur in one second. It is a measure of how many times the direction of the current reverses per second. Frequency is typically expressed in hertz (Hz), where one hertz equals one cycle per second. In practical terms, the frequency of AC current determines how rapidly the current alternates its direction and is crucial for the operation of electrical devices that rely on AC power.The relationship between frequency and time period is expressed as $f = \frac{1}{T}$

Instantaneous Peak Value:

The instantaneous peak value refers to the maximum amplitude of the alternating current or voltage at any specific point in time. For a sine wave, this is the highest positive or negative value reached during a cycle. It is usually denoted as I_{peak} for current or V_{peak} for voltage.

Root Mean Square (rms) Value:

The average value of voltage or current is not used in electric power calculations. The reason being that the AC cycle consists of positive and negative half cycles so the average value is zero. However, there exists a mathematical relation between the peak value V_0 of alternating voltage and a direct current (d.c) voltage that yields an equivalent average electrical power. This constant d.c. voltage is referred to as the root-mean-square (r.m.s) value of the alternating voltageV_{rms}. The d.c voltage is approximately 70% of V_{peak} as shown in figure 20.3. For a sinusoidal wave, the rms value is approximately 0.707 times the peak value. Mathematically,

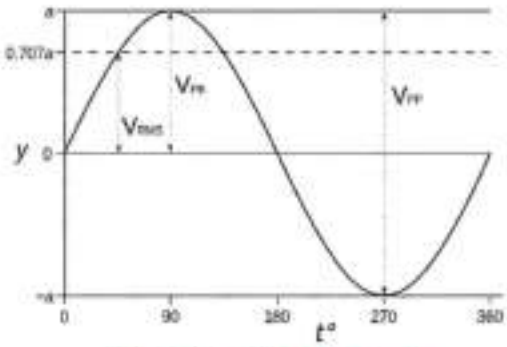


Figure 20.3 Cycle of ac current

$V_{rms} = 0.707 \times V_{peak}$ for voltage

and $I_{rms} = 0.707 \times I_{peak}$ for current.

20.1.2 Alternating Current OR voltage equation:

When a coil undergoes constant rotation at a uniform angular velocity within a magnetic field, it generates an alternating voltage. Alternating Current (AC) or Voltage refers to the type of electrical flow or potential difference that periodically changes direction and magnitude over time. The fundamental equation describing AC voltage or current in a sinusoidal waveform is:

$V = V_0 Sin\omega t$..... (20.1)

$I = I_0 Sin\omega t$..... (20.2)

Where,

V, I represent the instantaneous value of the voltage and alternating current respectively

V_0, I_0 is the amplitude of the waveform, which represents the maximum value of alternating voltage and alternating current respectively

ω is the angular frequency, given by $\omega = 2\pi f$

Worked Example 20.1

If an alternating current is flowing through a circuit with an amplitude of 5 amps, with frequency of 60 Hz, write its equation:

Solution:

Step 1: Write the known quantities and point out quantities to be found

$Vo = 5A$,

$f = 60\ Hz$

Step 2: Write the formula and rearrange if necessary

$$I = I_0 Sin\omega t$$

Step 3: Now, substitute the values into the formula:

$$I=5\ Sin2\pi\times60\times t$$

Result: The sine function will oscillate between -5 and 5 Amps.

20.1.3 Phase of AC:

The phase difference between two waves carries more important than their magnitudes. The phase describes the relative positions of these waveforms. Hence in alternating current (A.C.) circuits, the concept of phase refers to the relationship between different waveforms as time passes. Now we discuss different states of phase within AC circuit.

In-phase:

When two waveforms have the same frequency and reach their peak values or zero values at the same time, they are said to be in phase as shown in figure 20.4.

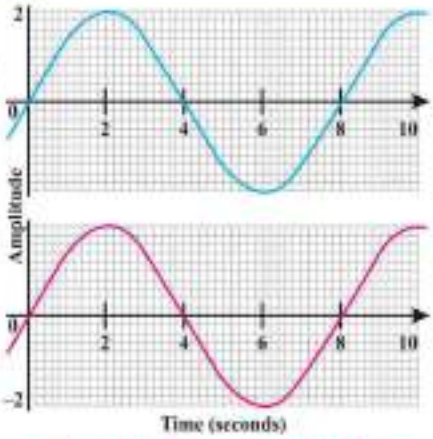


Figure 20.4 Two waves are in a phase

Phase Lag:

If one waveform reaches its peak or zero value after the other, there is a phase difference, and the second waveform is said to lag the first. It is shown by figure 20.5 where a wave shown by dotted line lags the wave shown by solid line. It can be said that the dotted wave form lags the solid line waveform by 90^0 or $\frac{\pi}{2}$ radians.

Phase Lead:

Conversely, if one waveform reaches its peak or zero value before the other, the second waveform is said to lead the first. It is shown by figure 20.5 where a wave shown by solid line leads the wave shown by dotted line by phase difference of 90^0 or $\frac{\pi}{2}$ radians.

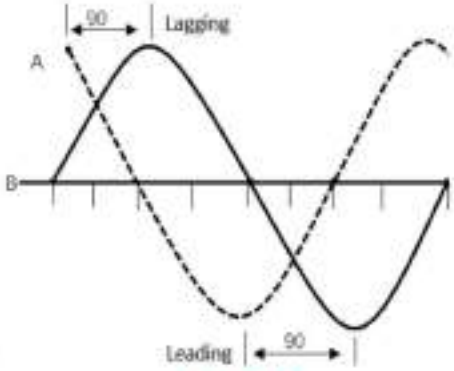


Figure 20.5
The two waves are 90^0 out of phase.

The representation of phase lead and lag between two alternating quantities is effectively demonstrated by showing the two AC quantities as vectors also known as "Phasor".

Vector Representation of an Alternating quantity:

To solve AC problems effectively, it is beneficial to represent a sinusoidal quantity, such as voltage or current, using a line of specific length that rotates in a counterclockwise direction with the same angular velocity as the sinusoidal quantity. This rotating line is commonly referred to as a phasor.

Let's consider a line denoted as OA, referred to as a phasor, which accurately represents, to scale, the maximum value of an alternating quantity, such as electromotive force (emf). In this representation, OA equals the maximum emf value and rotates counterclockwise at an angular velocity of ω radians per second around the point O, as illustrated in Figure 20.6. An arrowhead is placed at the outer end of the phasor, serving both to indicate the assumed direction of movement and to specify the precise length of the phasor, especially when multiple phasors coincide.

In Figure 20.6, OA represents the phasor after it has rotated through an angle θ, equivalent to ωt, from its initial position when the emf was at its zero value. The projection of OA on the Y-axis, denoted as OB, equals OA multiplied by sin θ, which is equivalent to Emax sin ωt, representing the instantaneous value of the emf, denoted as 'e,' at that particular moment. Therefore, the projection of OA on the vertical axis accurately portrays, to scale, the instantaneous value of the emf.

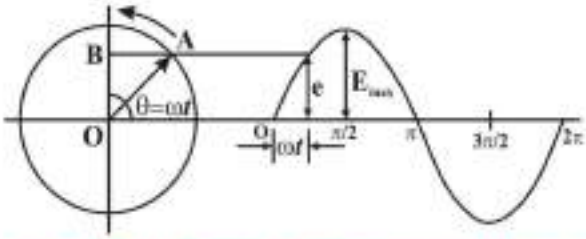


Figure 20.6 The propagation of sinusoidal wave and its vector (phasor) representation

Self-Assessment Questions:

1. Explain the concept of reactance in AC circuits and how it differs from resistance.
2. How does the phase angle affect the relationship between voltage and current in AC circuits?

20.2 AC through a resistor

Consider a circuit comprising a pure resistor (R) connected to an alternating voltage source, as depicted in figure 20.7(a).

The alternating voltage induces oscillatory motion of free electrons within the resistor, constituting the alternating current. At any given time 't,' the potential difference across the resistor's terminals is expressed as

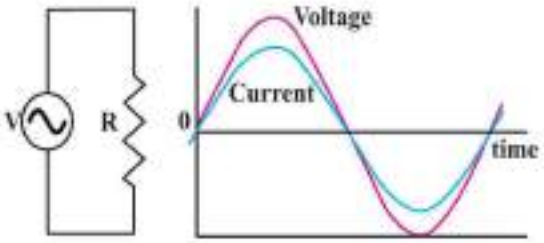


Figure 20.7(a) resistor connected to an AC source and 20.7(b)the voltage and current wave are in a phase

$$V = V_0 sin\omega t$$

where V_0 signifies the peak value of the alternating voltage. The circuit's current (I) is governed by Ohm's law: $I = \frac{V}{R} = \frac{V_0}{R} Sin\omega t$

Or $I = I_0 Sin\omega t$..... (20.3)

The current achieves its maximum value when $Sin\omega t = 1$. From equations (20.2) and (20.3), it is evident that the applied voltage and the current in the circuit are in phase with each other as shown in figure 20.7 (b). This phase relationship is further illustrated by the phasor diagram as shown in figure 20.8.

20.2.3 Phasor diagram of Resistive circuit

The current I_R flowing through a resistor R is in phase with the voltage V_R across the resistor. This alignment can be visually represented on a phasor diagram by depicting a vector (I_R) that coincides with the voltage (V_R)

The power loss in a resistor in an AC circuit is a result of the conversion of electrical energy into heat due to the resistance of the resistor. The instantaneous power in the resistance is can be expressed using the formula

$P = I^2 R$..... (20.4)

$P = V_{rms} I_{rms}$..... (20.5)

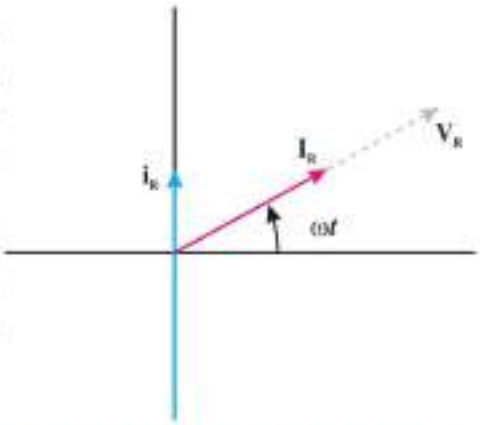


Figure 20.8 phasor diagram of resistive circuit

Worked Example 20.2

A 200Ω resistor is connected with AC supply of 36V with frequency 60 Hz. Calculate
(i) Current flowing through the circuit
(ii) Power dissipated in the resistor.
(iii) Also write equation for current and voltage as a function of time.

Solution:

Step 1: Write the known quantities and point out quantities to be found

$R = 20\Omega$
$V = 36\ Volts$
$f = 60\ Hz$

Step 2: Write the formula and rearrange if necessary

$$I = \frac{V}{R} = \frac{36}{200} = 0.180A$$

Step 3: Power dissipation can be calculate using equation

$P = VI = 36 \times 0.180 = 6.48$ w
We know that $\omega = 2\pi f = 2 \times 3.14 \times 60 = 376.8$ Hz
From equations (2) and (3)
$V = V_0 \sin\omega t$
$I = I_0 \sin\omega t$

Step 4: Substituting values of V, I and ω in above equations, we get

$V = 36 \sin 376t$
$I = 0.180 \sin 376t$

20.3.1 A.C through capacitors:

When capacitor is connected to a direct current (DC) supply voltage, then capacitor plates gradually accumulate charge until the voltage across the capacitor matches with the source voltage. The capacitor retains this charge as a temporary storage device as long as the applied voltage is sustained. The ability of a capacitor to store charge on its plates is termed capacitance (C)

During the charging process, an electric current flow into the capacitor, causing its plates to acquire an electrostatic charge. This charging current is highest when the capacitor plates are uncharged, decreasing exponentially over time until the capacitor reaches full charge. Once the capacitor is fully charged, it blocks further electron flow onto its plates as they become saturated. Hence it is often said that capacitor blocks direct current after initial transient time.

Now consider a capacitor connected to the AC source as shown in figure 20.9. during the positive half cycle of alternating voltage, the electrons flow from upper capacitor plate to

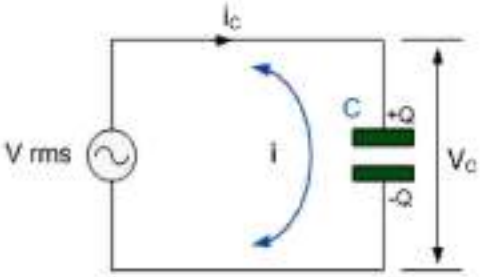


Figure 20.9 capacitor connected to the AC source

the source leaving it as positively charged +Q and source supplies equal number of electrons to the lower plate making it as negatively charged –Q as shown in figure 20.9. During the negative half cycle of alternating voltage, the direction of motion of electrons is also reverted resulting in the capacitor plates becoming charged in the opposite manner and current flows in opposite direction, hence capacitor charges and discharges. In this way the alternating current flows through the capacitor.

If alternating source voltage $V = V_0 sin\omega t$ is applied to the capacitor the charge on any plate of capacitor is given by

$q = CV = CV_0 Sin\omega t$..... (20.6)

From equation (20.6) it is clear that q and V are in phase. Due to the applied current flows through the capacitor as given by

$I = \frac{\Delta q}{\Delta t}$ ..... (20.7)

Substitute equation (20.6) in (20.7), we get

$I = \frac{CV_0 \Delta Sin\omega t}{\Delta t}$ OR $I = CV_0 \omega Cos\omega t$..... (20.8)

Multiply and divide right side of equation (20.8) by ωc

$I = \frac{V_0 Sin(\omega t + \frac{\pi}{2})}{1/\omega c}$

$I = I_0 \, Sin(\omega t + \frac{\pi}{2})$ ..... (20.9)

Equation (20.9) shows that in pure capacitive circuit the current exhibits sinusoidal variation and it leads the voltage by 90 degrees or $\frac{\pi}{2}$ radians or the voltage lags the current by 90 degrees or $\frac{\pi}{2}$ radians as shown in figure 20.10.

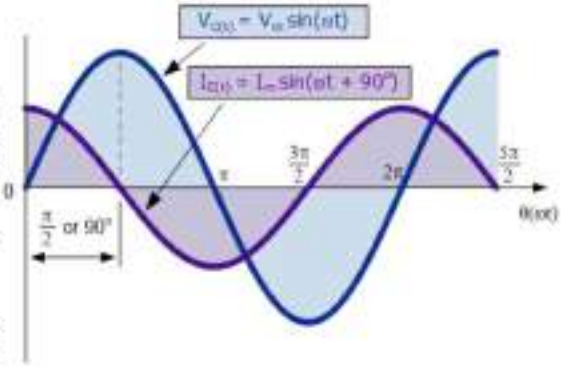


Figure 20.10 Pure capacitive circuit the current varies sinusoidally and leads the voltage by 90 degrees.

The reason is that when a voltage is introduced to an initially uncharged capacitor, the capacitor exhibits low impedance, resulting in the maximum current draw. As the capacitor becomes charged, the current diminishes, causing the voltage across the capacitor to increase. Once the capacitor is fully charged, there is no further current flow, and the voltage attains its maximum level, hence voltage lags the current by 90^0.

20.3.2 Reactance of Capacitor:

In a purely capacitive circuit, capacitive reactance represents the opposition to alternating current flow. Similar to resistance, reactance is measured in Ohms, but it is denoted by the symbol Xc to differentiate it from purely resistive values.

From equations (9), we get

$I = \frac{V_0}{1/\omega c}$..... (20.10)

Similar to ohms law the term in denominator of equation (20.10) represents the opposition to flow of current; in this case it is called reactance of capacitor.

Hence

$$X_c = \frac{1}{\omega c} \ldots\ldots\ldots\ldots (20.11)$$

Since $\omega = 2\pi f$ so

$$X_c = \frac{1}{2\pi f c}$$

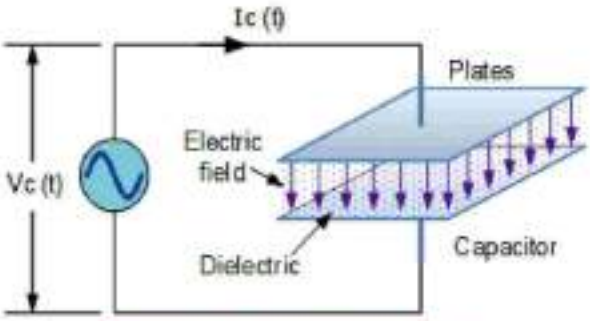


Figure 20.11 Reactance of Capacitor

The capacitive reactance shows an inverse relationship with the frequency of the applied alternating voltage. Consequently, for lower frequencies, the reactance of capacitor increases, while at high frequencies, the reactance decreases.

20.3.3 Phasor diagram of capacitive circuit:

As mentioned in section 20.3 that in an AC circuit, a capacitor experiences a phase shift between the voltage across it and the current flowing through it. The phasor diagram is a graphical tool that illustrates the relationship between voltage and current in a capacitor using vectors, known as phasor. In the phasor diagram shown in Figure 20.12, V represents the voltage across the capacitor, while I denote the current flowing through it. The angle $\emptyset = 90^o$ indicates the phase angle by which the current leads the voltage in a capacitor.

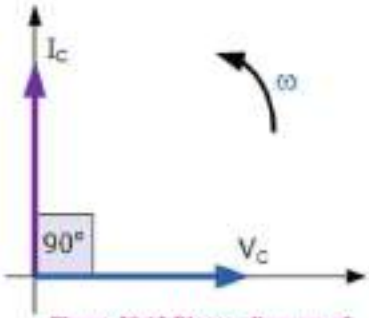


Figure 20.12 Phasor diagram of capacitive circuit

Self-Assessment Questions:

1. Why is the root mean square (RMS) value used to represent AC voltage and current instead of instantaneous peak values?
2. How does the phase angle affect the relationship between voltage and current in AC circuits?

Worked Example 20.3

A 200Ω resistor is connected with AC supply of 36V with frequency 60 Hz. Calculate
(i) Current flowing through the circuit (ii) Power dissipated in the resistor.
(iii) Also write equation for current and voltage as a function of time.

Solution:

Step 1: Write the known quantities and point out quantities to be found

$R = 200\Omega$
$V = 36\ Volts$
$f = 60\ Hz$

Step 2: Write the formula and rearrange if necessary

$$X_c = \frac{1}{2\pi fc}, \quad I_{rms} = \frac{V_{rms}}{X_c}$$

Step 3: Now, calculate the current: =

Xc =1/2πfc = 1/ (2×3.14×50×10×10 ^ (-6)) = 3183.1 Ω

I_{rms}= V_{rms}/X_c = 220/3183.1=0.069A

Step 3: First calculate the maximum current and voltage by using

I_0= √2I=0.097A

V_0=√2 V=311.1

V=311.1 Sin (314 t)

I=0.097 Sin (314 t+π/2)

20.4 A.C through inductor:

An inductor is a passive electrical component which can be formed by wounding a conducting wire over an insulating object, such as pencil. The primary purpose of an inductor is to oppose changes in current. It resists the flow of alternating current (AC) and stores energy in its magnetic field during the on-time of the AC cycle and releases it during the off-time. Solenoid is an example of inductor.

Consider an inductor in form of solenoid connected with A.C source as shown in figure 20.13.

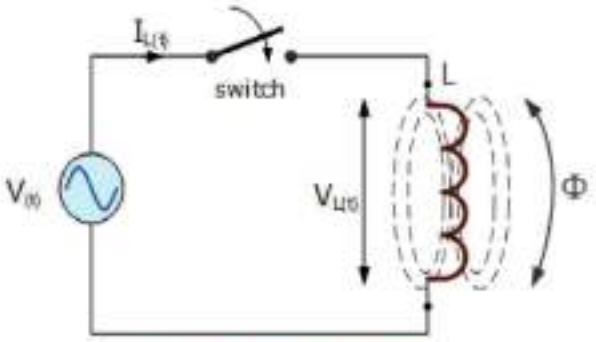


Figure 20. 13 A.C through inductor.

When the switch is closed current start to flow through the inductor, notice that the magnitude and direction of current is changing hence associated magnetic field also varies due to which an induced emf is set up in the inductor so as to oppose the change in accordance with the Lenz law. The magnitude of induced emf is given by

$$\varepsilon = L\frac{\Delta I}{\Delta t}$$

Therefore, to sustain the current, the applied voltage must match the back electromotive force (EMF). Thus, the magnitude of voltage supplied to the coil is expressed as

V= $L\frac{\Delta I}{\Delta t}$..... (20.12)

The alternating voltage produces a sinusoidal current given by

$$I = I_0 Sin\omega t$$

Hence equation (20.12) becomes

$V = L\frac{I_o \Delta sin\omega t}{\Delta t}$ Or

$$V = \omega L I_0 \, Cos\omega t$$

Where $\omega L I_0 = V_0$

$$V = V_0 \, Cos\omega t \ldots\ldots\ldots\ldots (20.13)$$

Using that $Cos\omega t = Sin\,(\omega t + \frac{\pi}{2})$, equation (13) becomes

$$V = V_0 \, Sin\,(\omega t + \frac{\pi}{2}) \ldots\ldots\ldots (20.14)$$

From equation (20.14) it is clear that in case of inductor connected in A.C circuit the voltage leads the current by $\frac{\pi}{2}$ radians as shown in figure 20.14.

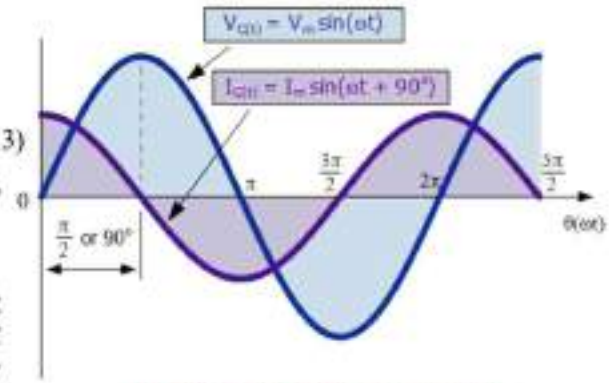


Figure 20.14 Voltage leads current by 90

The reason for voltage leading in an inductive circuit is due to the generation of induced electromotive force (EMF), commonly known as back EMF, when an alternating voltage is applied to the inductor. This back EMF appears instantaneously and induces a counter-current flow, introducing a time delay, typically in the order of milliseconds. As a result, there is a brief time lag for the circuit current to overcome this opposing current and attain its maximum value. That is why the voltage manifests first followed by the appearance of the current after a short interval. Hence the voltage leads the current or the current lags behind the voltage in an inductive AC circuit, draw I_L behind V_L by a phase angle of 90 degrees as shown by phasor diagram in figure 20.15.

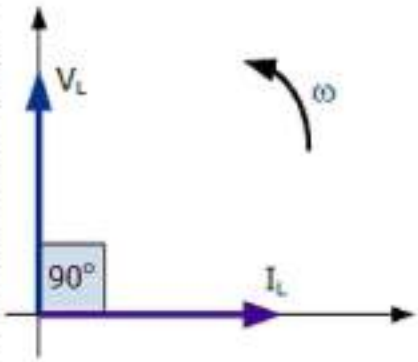


Figure 20.15
Phasor diagram of inductive circuit

Choke coil:

A choke coil, commonly known as choke, is an inductor used in electronic circuits. It is used to block high frequencies above a certain frequency, while allowing the direct current to pass through. It is formed within a core shaped like a donut, around which an insulated coil is wound as shown in figure 20.16. This important electrical component is only used in A.C circuits. Its ability to impede changes in current makes it valuable in applications where a smooth and noise-free DC signal is required. Choke coils find applications in various electronic devices, including power supplies for audio amplifiers, radio frequency (RF) circuits, and other electronic systems where filtering and noise suppression are essential.

Figure 20.16 Chock coil

20.4.1 Inductive Reactance:

The term inductive reactance refers to the opposition that an inductor offers to the flow of alternating current. It is denoted by X_L. It is measured in ohms. The inductive reactance can be calculated by following the method as described in section 20.3.1.

The formula for inductive reactance is given by:

$X_L = \omega L$ Or $\quad X_L = 2\pi f L$..... (20.15)

where: X_L is the inductive reactance,

f is the frequency of the AC signal,

L is the inductance of the inductor.

Equation (20.15) shows that inductive reactance is directly proportional to the frequency of the AC signal. As the frequency increases, the inductive reactance also increases.

Worked Example 20.4

Suppose we have an inductor in form of a coil with an inductance of 0.02henries (H) and connected in series with an AC signal having frequency of 50 Hz. Calculate the inductive reactance of coil.

Solution:

Step 1: Write the known quantities and point out quantities to be found

$L = 0.02H$

$f = 50Hz$

$X_L = ?$

Step 2: Write the formula and rearrange if necessary

$$X_L = 2\pi f L$$

Step 3: Substituting the give values in formula, we get

$X_L = 2 \times 3.14 \times 50 \times 0.02$

$X_L = 6.28\ \Omega$

20.5 RLC Circuits:

An RLC circuit is an electrical circuit consisting of a resistor (R), inductor (L), and capacitor (C), connected in series or parallel. RLC circuits are fundamental in electronics for filtering signals, tuning circuits, and resonant applications due to their ability to manipulate frequencies and impedance.

Impedance triangle:

A circuit containing resistor, inductor and capacitor offer opposition to flow of current due to all these circuit elements known as impedance. It is denoted by Z. The impedance triangle is shown in figure 20.17.

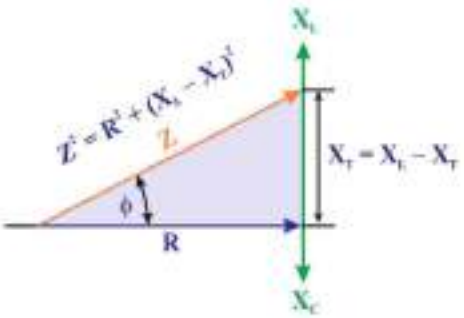


Figure 20.17 impedance triangle

To draw an impedance triangle, represent the resistance (R) as the horizontal side of the triangle.

Represent the inductive reactance (X_L) or capacitive reactance (X_C) as the vertical side of the triangle. The direction (up or down) depends on whether it's inductive or capacitive reactance.

Now the hypotenuse of the triangle represents the impedance (Z). From the right-angle triangle impedance can be calculated by using following relation

$Z = \sqrt{R^2 + X_T^2}$..... (20.16)

Where Z is impedance, R is resistance in a circuit and $X_T^2$is sum of capacitive and inductive reactance of circuit

20.5.1 AC through RC Series circuits:

Consider a circuit containing resistor and capacitor connected in series with an alternating voltage source as shown in figure 20. 16. Due to series circuit same current will flow through each circuit element. While voltage across resistor will be V_R and V_c across capacitor.

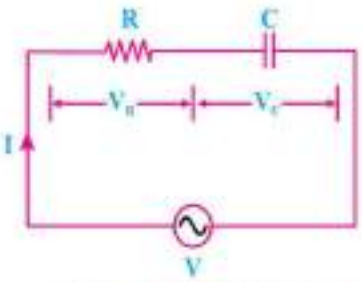


Figure 20.18 RC series circuit

The phasor diagram is drawn taking current as reference direction as shown in figure 20.18. As discussed in section 20.3.3 the voltage lags the current by $\frac{\pi}{2}$ radians in case of capacitor so it is drawn perpendicular to the current phasor in figure 20.19. In RC series circuit the voltage and current are in phase when considering resistor component of circuit.

From the phasor diagram, using pythagoras theorem

$$V = \sqrt{V_R^2 + V_C^2} \text{..... (20.17)}$$

Substituting volatage across resistor = IR and across capacitor =IXc in eqaution (16), we get

$$V = I\sqrt{R^2 + X_C^2} \text{..... (20.18)}$$

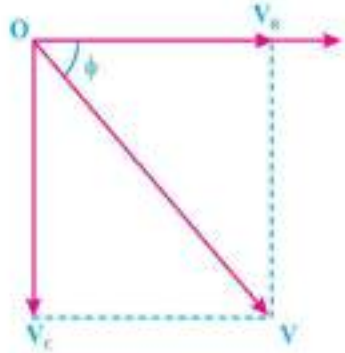


Figure 20.19
Phasor diagram of RC series circuit

Comparing eqautions (20.16) and (20.18), we find that qauntity $\sqrt{R^2 + X_C^2}$ represents the impedence of RC series circuit. $Z = \sqrt{R^2 + X_C^2}$

Phase angle between current and voltage as shown in figure 20.19 is calculated by

$$tan\emptyset = \frac{V_C}{V_R} = \frac{X_C}{R} \qquad \text{Or} \quad \emptyset = tan^{-1}(\frac{X_C}{R})$$

20.5.2 AC through RL Series circuits:

Next in order to calculate the impedence of RL series circuit, consider a circuit containing indcutor and resistor connected in series with an alteranting voltage as shown in figure 20.20(a). The volatge across resistor and indcuitor in V_R and V_L respectively.

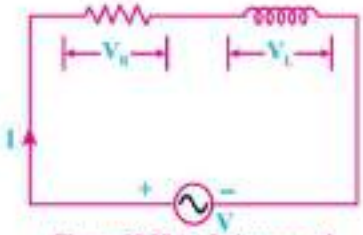


Figure 20.20 (a) Inductor and resistor connected in series

As discussed in section 20.3.6 when an indcutor is connected with alternating voltage then due to generation of back EMF current lag the voltage by 90^0 it is repredented by line perpendciular to the referenece phasor i-e current I as shown in fgiure 20.20(b). Also as discussed in section 20.31 in resisitor the current and voltage are in phase. In phasor diagram it is V_R line.

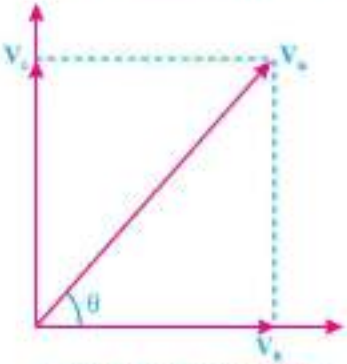


Figure 20.20 (a) Inductor and resistor connected in series

To calculate the impedance of RL circuit, apply pythagoras theorem on phasor diagram from figure 20.18 (b)

$$V = I\sqrt{R^2 + L_C^2}$$

Where $Z = \sqrt{R^2 + L_C^2}$ represents the impedance of the RL circuit.

Phase angle between current and voltage as shown in figure 20.20(b) is calculated by

$tan\emptyset = \frac{X_L}{R}$ Or $\emptyset = tan^{-1}\left(\frac{X_L}{R}\right)$

Worked Example 20.5

In RC circuit a capacitor of 50μF and resistor of 100 Ω are connected in series with alternating voltage source of 220V and frequency of 50Hz. Find the (a) current in the circuit (b) phase angle between voltage and current.

Solution:

Step 1: Write the known quantities and point out quantities to be found

$C = 50\times10^6F$ $R = 100\ \Omega,$ $V = 220Volts,$ $f = 50Hz$

Step 2: Write the formula and rearrange if necessary

$$X_c = \frac{1}{2\pi fc} =$$

Step 3: Substituting the give values in formula, we get

The capacitive reactance $= X_c = \frac{1}{2\pi fc} = \frac{1}{6.28\times50\times50\times10^{-6}} = 636.9H$

Impedance $= Z = \sqrt{R^2 + X_c^2} = \sqrt{100^2 + 636.9^2} = 644.7\Omega$

Current $= I_{rms} = \frac{V_{rms}}{Z} = \frac{220}{636.9} = 345mA$

Phase angle $= \emptyset = tan^{-1}\left(\frac{X_c}{R}\right) = tan^{-1}\left(\frac{636.9}{100}\right) = 81.04^0$

20.5.3 RLC Series AC Circuit:

An RLC series circuit consists of a resistor (R), inductor (L), and capacitor (C) connected in series to an AC voltage source as shown in figure 20.21. In such a circuit, the electrical components influence the behavior of the current flowing through the circuit.

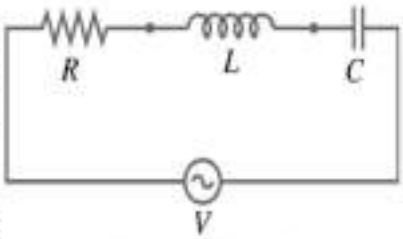


Figure 20.21 RLC Series AC circuit

Since all three circuit elements are connected in series so same current will flow through them. Hence to analyze the different properties of circuit we draw a phasor diagram of RLC series circuit as shown in figure 20.22.

In section 20.4 and 20.5 the capacitive and reactance were determined as $X_c = \frac{1}{2\pi fc}$ and $X_L = 2\pi fL$. for smaller frequency capacitive reactance is much greater than inductive reactance i-e $X_C > X_L$ causing the circuit to exhibit behavior akin to an RC circuit. For larger frequency inductive reactance dominates over capacitive reactance i-e $X_L > X_C$ and circuit behaves like an RL circuit. Hence the circuit is said to be more inductive than capacitive.

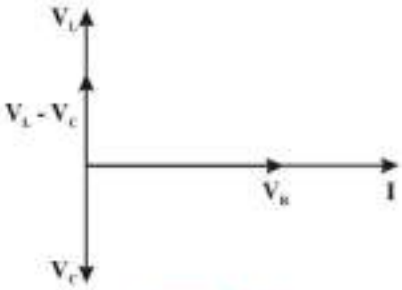


Figure 20.22
Phasor diagram of RLC series circuit

20.5.4 Power in A.C Circuits:

The calculation of power dissipation in a resistor involves the formula $P = VI$. However, this equation is not applicable in circuits where capacitors and inductors are connected to an alternating voltage source. The reason being that in a purely resistive circuit, the current and voltage are in phase. Whereas circuits containing capacitive or inductive elements exhibit a lead or lag of $\frac{\pi}{2}$ radians between the current and voltage, respectively, disrupting the applicability of this relationship. Therefore, in such a cases to determine power dissipation, project V onto the direction of referenece current phasor as shown in figure 20.18 (b). This approach ensures that the applied voltage is in phase with the current. Hence power dissipation in AC circuit is expressed as follows

$$P = VIcos\theta$$

where$cos\theta$ represents power factor.

Worked Example 20.6

A resistor (R) of 10 Ω and an inductor (L) of 0.05 H are connected in series. The AC voltage source has a peak voltage of 100 V and a frequency of 50 Hz.

Find:

1. The impedance of the circuit.
2. The RMS current through the circuit.
3. The average power consumed by the circuit

Solution:

Step 1: Write the known quantities and point out quantities to be found

$R = 10\ \Omega$

$L = 0.05$ H

$V_{peak} = 100$V

$f = 50$ Hz.

Step 1: Impedance of the Circuit (Z):

$$X_L = \omega_L = 2\pi f L = 2\pi \times 50 \times 0.05 = 15.7\ \Omega$$

$$Z = \sqrt{R^2 + X_L^2} = \sqrt{10^2 + 15.7^2} = \sqrt{100 + 246.49} = \sqrt{346.49} = 18.6\Omega$$

Step 2: RMS Current (I):

$$V_{RMS} = \frac{V_{peak}}{\sqrt{2}} = \frac{100}{\sqrt{2}} = 70.7V$$

$$I_{RMS} = \frac{V_{RMS}}{\sqrt{2}} = \frac{70.7}{18.6} = 3.8A$$

Step 3: Average Power (P):

$$P = V_{RMS} \times I_{RMS} \times \cos\phi$$

$$\text{Where } \cos\phi = \frac{R}{Z} = \frac{10}{18.6} = 0.54$$

$$P = 70.7 \times 3.8 \times 0.54 = 145W$$

Self-Assessment Questions:

1. Compare and contrast the flow of AC through resistors versus capacitors in terms of phase shift.
2. Describe the conditions under which maximum power transfer occurs between a source and a load in an AC circuit.

20.5.5 Resonant Frequency:

For a certain frequency the capacitive and inductive reactance becomes equal, $X_C = X_L$. This frequency is called resonant frequency f_r and circuit is said to be in resonance state. To calculate the resonant frequency, use the fact that

$$X_C = X_L$$

$$\frac{1}{2\pi fc} = 2\pi f L \text{ or} f_r = \frac{1}{2\pi\sqrt{LC}} \ldots\ldots (20.19)$$

From figure 20.23 note that X_C *and* X_L are in opposite direction. Therefore, at resonant frequency they cancel each effect in circuit. Now opposition to the current flow is solely offered by

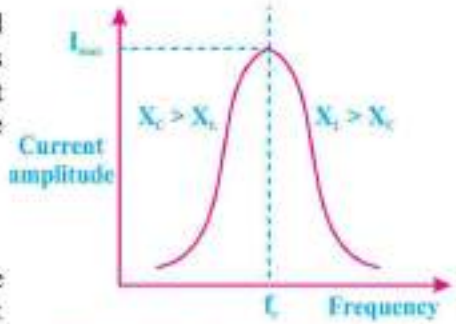


Figure 20.23 the resonant frequency

resistor, resulting in maximum current to flow through the circuit, moreover either side of resonant frequency the current in the circuit decreases, as shown in figure 20.23. RLC series circuit is an important circuit it is used in; Electronic filters, Electromagnetic Interference (EMI) Suppression, Tuned Circuits, etc

20.5.6 Parallel RL Circuit:

In some respects the circuit of figure 20.24 is similar to the purely inductive parallel circuit.

For instance, applied voltage V is still the quantity which is common to both components and is therefore plotted in standard position in the phasor diagram. Also the magnitude of the individual branch currents is determined by the opposition (reactance) of the individual branches. The figure 20.25 shows a composite diagram of waveforms and phasors. Since the phasor diagram shows that the two branch currents are not in phase, it will be necessary to use phasor addition in order to determine the total current.

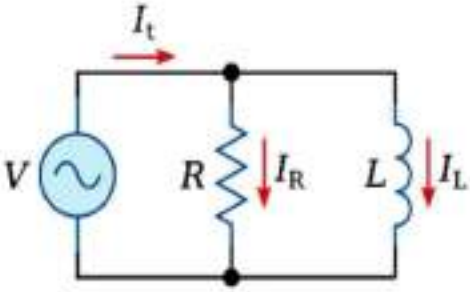


Figure 20.24 Parallel RL Circuit

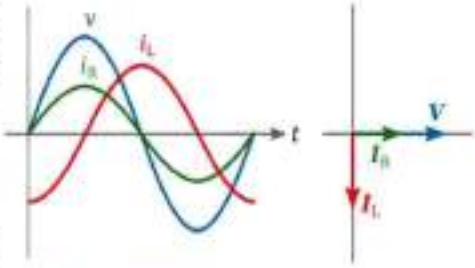


Figure 20.25 Waveforms and phasors

20.5.7 Parallel RC Circuit:

Parallel RC circuits may be resolved in much the same way as are parallel RL circuits. The figure 20.26 illustrates a parallel RC circuit.

The figure 20.27 shows a composite diagram of waveforms and phasors as per circuit conditions. The current phasors IR and IC are out of phase; therefore, phasor addition must be used to determine total current. The solving of an RC circuit follows the method previously applied to LR circuits.

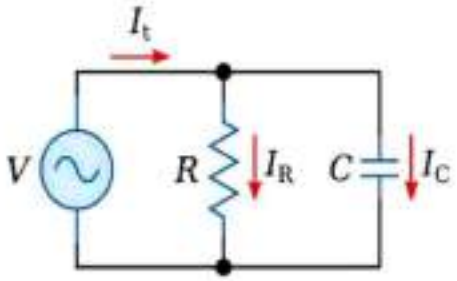


Figure 20.26 Parallel RC Circuit

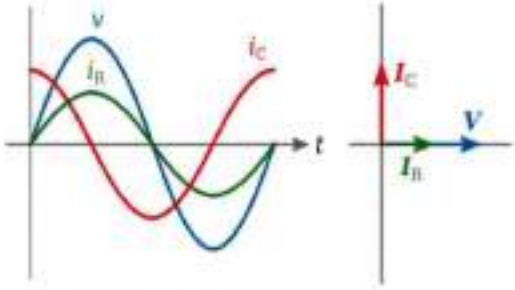


Figure 20.27 Waveforms and phasors

20.5.8 Parallel RLC AC Circuit:

A Parallel RLC AC Circuit is one where the resistor, inductor, and capacitor are connected in parallel to each other and the AC source. In this configuration, the voltage across each component is the same, but the currents through them differ. The parallel arrangement affects the overall impedance and current distribution in the circuit, making it distinct from its series counterpart. In a parallel circuit, the voltage V (RMS) across each of the three elements remains the same. Hence, for convenience, the voltage may be taken as a reference phasor.

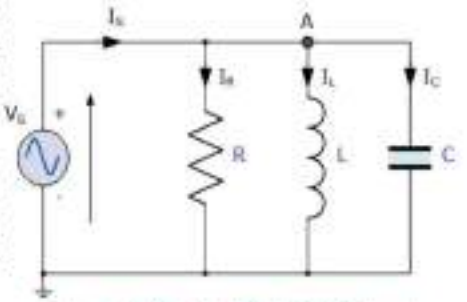


Figure 20.28 Parallel RLC AC Circuit

In the parallel RLC circuit, the supply voltage V_S is common to all three components, while the supply current I_S consists of three parts: the current through the resistor (I_R), the current through the inductor (I_L), and the current through the capacitor (I_C). The current flowing through each branch, and therefore through each component, will differ from one another and from the supply current I_S.

- The total current drawn from the supply is not simply the arithmetic sum of the three individual branch currents but their vector sum.
- Similar to the series RLC circuit, we can solve this circuit using the phasor or vector method. However, in this case, the vector diagram will use the voltage as its reference, with the three current vectors plotted relative to this reference voltage.
- The phasor diagram for an AC RLC parallel circuit is created by combining the individual phasors for each component and adding the currents vectorially.

Since the voltage across the circuit is common to all three circuit elements, we can use it as the reference vector, with the three current vectors drawn relative to this reference at their corresponding angles. The resulting vector current I_S is obtained by first adding the vectors I_L And I_C, and then adding this sum to the vector I_R. The angle between V and I_S.

20.5.9 Phasor Diagram of a Parallel RLC Circuit:

From the phasor diagram of AC RLC parallel circuit given below, we observe that the current vectors form a right triangle, with the hypotenuse represented by I_S, the horizontal axis by I_R, and the vertical axis by I_L−I_C. This configuration forms what is known as a Current Triangle. Consequently, we can apply Pythagoras's theorem to this current triangle to mathematically determine the individual magnitudes of the branch currents along the x-axis and y-axis. This will allow us to calculate the total supply current I_S of these components, as illustrated in figure. The circuit's phase angle is give.

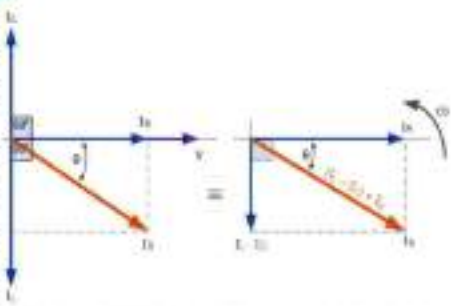


Figure 20.29 Phasor Diagram of a Parallel RLC Circuit

20.5.10 Resonance of Parallel RLC AC Circuit:

A parallel circuit containing a resistance, R, an inductance, L and a capacitance, C will produce a parallel resonance (also called anti-resonance) circuit when the resultant current through the parallel combination is in phase with the supply voltage. At resonance there will be a large circulating current between the inductor and the capacitor due to the energy of the oscillations, then parallel circuits produce current resonance.

A parallel resonant circuit stores the circuit energy in the magnetic field of the inductor and the electric field of the capacitor. This energy is constantly being transferred back and forth between the inductor and the capacitor which results in zero current and energy being drawn from the supply.

In the solution of AC parallel resonance circuits we know that the supply voltage is common for all branches, so this can be taken as our reference vector. Each parallel branch must be treated separately as with series circuits so that the total supply current taken by the parallel circuit is the vector addition of the individual branch currents.

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

$$I_L = \frac{V}{X_L} = \frac{V}{2\pi fL}$$

$$I_C = \frac{V}{X_C} = V.2\pi fC$$

Therefore, I_T = vector sum of $(I_R + I_L + I_C)$

$$I_T = \sqrt{{I_R}^2 + (I_L + I_C)^2}$$

At resonance, currents IL and IC are equal and cancelling giving a net reactive current equal to zero. Then at resonance the above equation becomes.

$$I_T = \sqrt{{I_R}^2 + (0)^2} = I_R$$

Therefore, the circuit current at this frequency will be at its minimum value of V/R and the graph of current against frequency for a parallel resonance circuit is shown in figure.

The frequency response curve of a parallel resonance circuit shows that the magnitude of the current is a function of frequency and plotting this onto a graph shows us that the response starts at its maximum value, reaches its minimum value at the resonance frequency when IMIN = IR and then increases again to maximum as f becomes infinite.

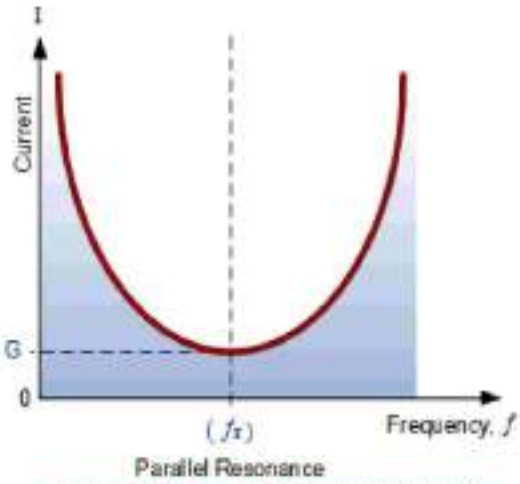


Figure 20.30 The graph of current against the frequency of parallel resonance circuit

The result of this is that the magnitude of the current flowing through the inductor, L and the capacitor, C tank circuit can become many times larger than the supply current, even at resonance but as they are equal and at opposition (180° out-of-phase) they effectively cancel each other out.

As a parallel resonance circuit only functions on resonant frequency, this type of circuit is also known as an **Rejecter Circuit** because at resonance, the impedance of the circuit is at its maximum thereby suppressing or rejecting the current whose frequency is equal to its resonant frequency. The effect of resonance in a parallel circuit is also called "current resonance".

Self-Assessment Questions:

1. What factors determine the resonant frequency of an LC circuit?
2. How does the quality factor (Q-factor) relate to the sharpness of resonance in a circuit?

Worked Example 20.7

An inductor in for of solenoid have reactance of 500Ω is connected with alternating voltage 220V and frequency of 50 Hz. Calculate the inductive reactance of the solenoid.

Solution:

Step 1: Write the known quantities and point out quantities to be found

$V = 220V,$
$f = 50\ Hz,$
$X_L = 500\Omega$
$Find\ L = ?$

Step 2: Write the formula and rearrange if necessary

$$X_L = 2\pi f L$$

Step 3: Substituting the give values in formula, we get

$$L = \frac{X_L}{2\pi f} = \frac{500}{6.28 \times 50} = 1.59 Henry$$

20.5.11 Maximum Power Transfer Theorem:

According to this theorem, maximum power is transferred from a source to a load when the impedance of the source matches with the impedance of the load.

In previous sections we have discussed that impedance is a measure of the opposition a circuit presents to the flow of alternating current (AC). Impedance includes both resistance and reactance, where reactance is the opposition due to capacitance or inductance.

When the impedance of the source matches the impedance of the load, the conditions for maximum power transfer are met. Under these circumstances, the electrical energy is efficiently transferred from the source to the load, resulting in the maximum power transfer.

Mathematically, suppose R_L denotes the resistance of load and R_s is source resistance.

Since $R_L = R_s$ as shown in figure 20.31 so maximum power transfer will occur.

In practical terms, achieving maximum power transfer is often important in designing electronic circuits to optimize efficiency and minimize power loss.

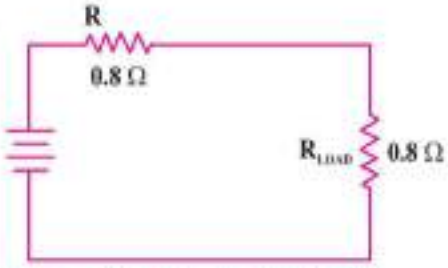


Figure 20.31 maximum power transfer

Worked Example 20.8

For an RLC circuit in parallel with a 5 Ω resistor, a 0.02 H inductor, and a 0.005 F capacitor, calculate the resonant frequency f_r?

Solution:

Step 1: Write the known quantities and point out quantities to be found

$L=0.02H$

$C=0.005F$

$f_r = ?$

Step 2: Write the formula and rearrange if necessary

$$f_r = \frac{1}{2\pi\sqrt{LC}}$$

Step 3: Substituting the give values in formula, we get

$$f_r = \frac{1}{2\pi\sqrt{0.02 \times 0.005}}$$

$$f_r = 15.92\ Hz$$

20.6.1 Metal Detectors:

The oscillator circuit is used in metal detectors. Metal detectors used for security checks operate on the principle of electromagnetic induction. A simplified explanation of working of metal detector is described below:

1. **Generating an Electromagnetic Field:**
   The metal detector contains a coil of wire through which an electric current flow, creating a magnetic field around the coil. This coil is often housed in a special arrangement, such as a loop or wand.
2. **Interaction with Metals:**
   When a conductive metal object is brought into the vicinity of the electromagnetic field, it disturbs the field. This disturbance induces a secondary magnetic field in the metal object.

3. **Eddy Currents:**
   The changing magnetic field induces circulating electric currents within the metal object, known as eddy currents. These eddy currents, in turn, generate their own magnetic fields.
4. **Detection of Changes:**
   The metal detector has a receiver coil or coils that are in close proximity to the transmitter coil. The receiver coil(s) detect changes in the magnetic field caused by the presence of the metal.
5. **Alert Mechanism:**
   When the metal detector senses a significant change in the magnetic field, indicating the presence of a metal object, it triggers an alert. This alert can be in the form of an audible sound, a visual signal, or both, depending on the design of the metal detector.

20.6.2 The Electrocardiograph:

The electrocardiograph (ECG) is a medical diagnostic tool used to record the electrical activity of the heart over a period of time.

Principle:

The principle behind the electrocardiograph is based on the electrical signals generated by the heart during each heartbeat.

Working:

The heart is a muscular organ that contracts rhythmically to pump blood throughout the body. This contraction is initiated and coordinated by electrical signals generated within the heart. Electrodes are attached to specific points on the skin of the patient. The standard placement involves attaching electrodes to the limbs and chest. These electrodes are conductive and are used to detect the electrical signals produced by the heart. The electrical signals produced by the heart are picked up by the electrodes. The detected electrical signals are amplified to make them more measurable and are then recorded on a graph or displayed on a monitor. The resulting graph is called an electrocardiogram. The ECG graph represents the electrical activity of the heart over time. It consists of waves and intervals, each of which corresponds to a specific phase of the cardiac cycle. Physicians analyze the ECG to gather information about the heart's rhythm, rate, and various other aspects of its electrical activity. Deviations from the normal ECG pattern can indicate cardiac abnormalities, such as arrhythmias, ischemia, or other heart conditions.

20.6.3 Oscillator Circuit:

An oscillator circuit is an electronic circuit that generates a continuous periodic signal at a specific frequency.

Oscillator circuits produce the carrier wave, which serves as the central frequency around which the information-carrying signal is modulated.

Oscillator circuits are designed to be tunable, allowing broadcasters to set the carrier frequency to a specific value. This turnability is essential for assigning unique frequencies to different radio stations, preventing interference between them. Oscillators are designed to be

modulated by an information-carrying signal. In amplitude modulation, the amplitude of the carrier wave is varied according to the audio signal, while in frequency modulation; the frequency of the carrier wave is modulated. This modulation process allows the transmission of audio information. The carrier wave itself does not convey information; it serves as a medium to carry the modulated information signal. The oscillator generates a stable carrier wave, ensuring that the modulated signal can be reliably transmitted over long distances. The carrier wave, when modulated with the audio signal, forms the composite radio wave that propagates through space. This transmitted signal can be received by radio receivers tuned to the carrier frequency.

In summary, the oscillator circuit is the heartbeat of a radio transmitter, generating the carrier wave that facilitates the transmission of modulated information signals. Its stability, tunability, and modulation capabilities are pivotal for effective and interference-free radio broadcasting.

20.6.4 Resonance in Tuning circuit of Radio:

In radio tuning circuits resonance is a important phenomenon that enables the selective reception of a desired radio frequency while rejecting others.

Principle:

The principle of resonance is crucial in the tuning circuits of radio to enhance the reception of specific radio frequencies. Resonance occurs when the inductive and capacitive reactance in a circuit cancel each other out, resulting in a condition where the circuit efficiently absorbs energy at a particular frequency. Resonance is integral to the functioning of radio tuning circuits. It ensures that radios can selectively receive signals from desired stations with clarity and efficiency, making it possible to enjoy various broadcasted content without interference from other frequencies.

Importance of Broadcasting:

Broadcasting remains a powerful medium with far-reaching impacts on society. It informs, educates, entertains, and connects people, playing a crucial role in cultural preservation, economic development, and social cohesion. Its ability to reach a wide audience makes it an indispensable tool for communication in the modern world.

SUMMARY

- AC or alternating current refers to the electric charge flow that periodically reverses direction, typically sinusoidal.
- The frequency of AC current, determines how many times per second the AC current reverses direction.
- Root Mean Square (RMS equivalent steady DC voltage producing the same power in a resistive load as the AC voltage.
- Current and Voltage are in phase in a resistive load known as Ohm's Law for AC current as $I = \frac{V_{RMS}}{R}$
- In pure capacitive circuit the current leads the voltage by 90 degrees or $\frac{\pi}{2}$ radians
- In inductive circuit the current lags the voltage by 90 degrees or $\frac{\pi}{2}$ radians
- The combined effect of resistance and reactance in RC or RLC circuit is known as impedance
- Resonant frequency is the frequency at which an oscillating system naturally vibrates with the greatest amplitude. At resonant frequency the impedance of the system is at its minimum.
- The RLC series circuit configuration results in high impedance at resonance and a lower overall current compared to the individual component currents.
- The RLC parallel circuit configuration the components are connected in parallel branches, providing multiple current paths. At resonance, the impedance is at its minimum, allowing a higher overall current flow compared to the series circuit.
- Maximum power is transferred from a source to a load when the load impedance is equal to the source impedance.
- ECG is a medical device that records the electrical activity of the heart over time.
- Oscillator Circuit is an electronic circuit that produces a periodic, oscillating signal, often a sine wave or square wave.
- Choke Coil is an inductor used to block higher-frequency AC signals while allowing lower-frequency or DC signals to pass.
- Resonance is the condition in a circuit when the inductive reactance and capacitive reactance are equal in magnitude but opposite in phase, resulting in a maximum voltage or current at a particular frequency.

EXERCISE

Section (A): Multiple Choice Questions (MCQs)

Choose the correct answer:

1. In alternating current, the flow of electric charge periodically reverses direction. The term for this specific phenomenon is:
   (a) Alternation (b) Oscillation (c) Cycle (d) Resistance
2. The frequency of standard household alternating current in Pakistan is:
   (a) 50 Hz (b) 60 Hz (c) 100 Hz (d) 120 Hz
3. The parameter for measuring the rate of change of alternating current or voltage with respect to time is:
   (a) Amplitude (b) Frequency (c) Phase angle (d) Peak Voltage
4. In AC circuits, what is the phase difference between current and voltage in a purely resistive circuit?
   (a) 0^0 (b) 45^0 (c) 90^0 (d) 180^0
5. A device used to measure the root mean square (RMS) value of an AC voltage or current is:
   (a) Voltmeter (b) Oscilloscope (b) Ammeter (d) AVO meter
6. In an AC circuit. The term "reactance" represent the:
   (a) Ohmic resistance only
   (b) Resistance to DC currents only
   (c) Capacitive and inductive opposition to current flow
   (d) Total impedance, including resistance and phase angle
7. A type of a circuit characterized by equal values of resistance and reactance is:
   (a) Resistive (b) Inductive (c) Capacitive (d) Resonant
8. In an AC sine wave, the peak-to-peak voltage is:
   (a) Double the peak voltage
   (b) The difference between the peak voltage and zero
   (c) The sum of the peak voltage and zero
   (d) The difference between the positive and negative peak voltages
9. What is the power factor of a purely resistive circuit?
   (a) 0 (b) 0.5 (c) 1 (d) -1
10. The purpose of a transformer in an AC power distribution system.
    (a) To convert AC to DC (b) To step up or step down voltage
    (c) To store electrical energy (d) To regulate current flow

Section (B): CRQs (Short Answered Questions):

1. Define RMS voltage and explain its significance in AC circuits.
2. Explain the difference between peak voltage, RMS voltage, and average voltage in AC circuits.
3. Define alternating current (AC) and explain how it differs from direct current (DC).
4. Define reactance in AC circuits and differentiate between capacitive and inductive reactance.

5. Describe the behavior of capacitors and inductors in AC circuits.
6. Explain the phase relationship between voltage and current in capacitive and inductive AC loads.
7. What is resonance in AC circuits? Discuss its conditions and applications.
8. Discuss the role of transformers in AC circuits and explain how they work.
9. Explain the behavior of an RLC series circuit in terms of impedance, resonance, and phase angle.
10. Describe the operation and applications of a transformer in AC circuits.

Section (C): ERQs (Long Answered Questions):

1. Explain the concept of phasor in the context of AC voltage. How are phasors used to represent sinusoidal voltages?
2. Describe the concept of impedance in AC circuits. How does impedance differ from resistance, and what are the units of impedance?
3. How does the reactance of an inductor and a capacitor change with frequency in an AC circuit? Provide an explanation based on the fundamental formulas?
4. Discuss the concept of resonance in RLC circuits. What conditions lead to resonance, and how does it affect the behavior of the circuit?
5. Compare and contrast series and parallel resonance in RLC circuits. What are the key differences between the two resonance configurations?
6. Explain the transient response of an RLC circuit when initially connected to an AC source. What happens to the currents and voltages in the circuit?
7. Why alternating current (AC) is commonly used for long-distance power transmission?
8. Describe the purpose and function of a choke coil in AC circuits.

Section (D): Numerical:

1. A resistor (R) of 20 ohms is connected in series with a capacitor (C) of 10 μF in an AC circuit with a frequency of 50 Hz. Calculate the total impedance? **(3182 ohms)**
2. For an inductor with an inductance (L) of 0.5 H and a frequency of 100 Hz, calculate the inductive reactance? **(314.16 ohms)**
3. In an RL circuit, the resistance (R) is 30 ohms, and the inductance (L) is 0.2 H. Calculate the total impedance at a frequency of 60 Hz? **(81.1ohms)**
4. In an RC circuit, the resistance (R) is 50 ohms, and the capacitance (C) is 20 μF. Calculate the capacitive reactance? **(31.83 ohms)**
5. An AC circuit has a resistance of 40 ohms, an inductive reactance of 30 ohms, and a capacitive reactance of 20 ohms. Draw the impedance triangle and calculate the total impedance? **(41.2ohms)**
6. In a series RL circuit, the resistance (R) is 25 ohms, and the inductance (L) is 0.1 H. Calculate the phase angle and impedance at a frequency of 80 Hz? **(63.8^0, 502 ohms)**
7. In a parallel RC circuit, the resistance (R) is 60 ohms, and the capacitance (C) is 30 μF. Calculate the total current flowing through the circuit at a frequency of 120 Hz? **(60 V)**
8. In an RLC circuit, the resistance (R) is 50 ohms, the inductance (L) is 0.1 H, and the capacitance (C) is 50 μF. Calculate the resonance frequency? **(70.7Hz)**

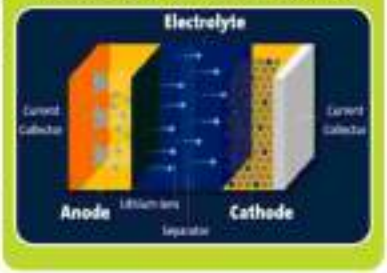
Introduction:

The Physics of solids is a branch of physics that deals with the study of the physical properties of solid materials, including their electronic, magnetic, optical, thermal and mechanical properties. It aims to understand the behavior of solid materials under different conditions, such as temperature, pressure and electromagnetic fields, using principles of quantum mechanics.It also helps to explain phenomenon like crystal structures, phase transitions, defects in material, and the interactions between electrons, ions and atoms in solids.

For example, a key hurdle to enhance solar power which is an affordable energy source. It is increasing the efficiency of photovoltaic cells that convert solar radiation to electricity, making solid materials in solar cells as conductive as possible

DO YOU KNOW?

John B, who credited for the identification and initial development of the first lithium-ion (Li-ion) batteries. He realized how a battery with lithium anode could provide high charge density and thus give us the pre-requisite technology for electric cars, and smart phones.



21.1 Classification of Solids:

Solids are usually crystalline substances their molecules are arranged in a definite pattern and in fixed positions.The way in which a solid behaves depends on its internal structure and there are three types of solid.

(i) Crystalline solids (ii) Amorphous solids (iii) Polymeric solids

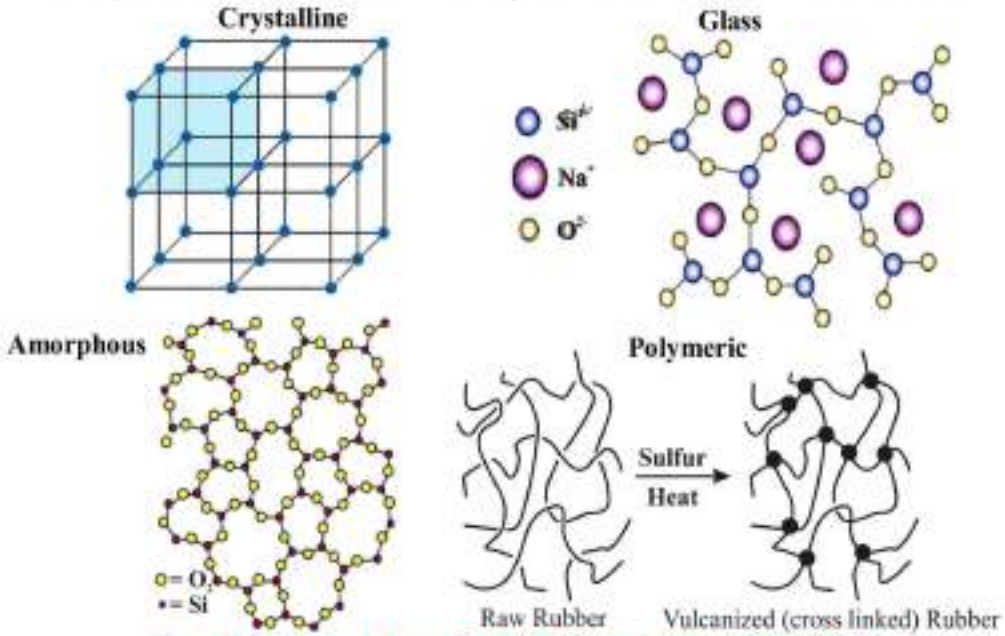


Figure 21.1 structures of crystalline, glassy, amorphous and polymeric solids

21.1.1 Distinguish between structure of crystalline, glassy, amorphous and polymeric solids:

Property	Crystalline Solids	Glassy Solids	Amorphous Solids	Polymeric Solids
Atomic Arrangement	Regular, Repeating pattern.	Random, short-range order.	Random, no long-range order.	Long chains or networks of repeating units.
Melting Point	Sharp, Specific temperature.	Gradual, over a range of temperature.	Gradual, over a range of temperatures.	Varies, typically lower than crystalline solids.
Rigidity	Very rigid and well-defined structure.	Rigid, but less than crystalline solids.	Not as rigid as crystalline solids.	Can be flexible or rigid.
Transparency	Can be transparent or translucent.	Transparent or opaque depending on composition.	Opaque.	Depending on structures, can be transparent or opaque.
Examples	Diamond, salt crystals, silicon.	Window glass, certain plastics, amorphous metals.	Rubber, some plastics, glass.	Polyethylene, PVC, nylon.

21.1.1 Deformation:

In material science, deformations refer to modifications of the shape or size of an objectdue to applied forces or a change in temperature. Deformation is usually caused by forces such as tensile (pulling) or Compressive (pushing) as shown in figure 21.2. As deformations occurs, internal inter-molecular forces arise that oppose the applied force.Different types of deformations may result from variations in type of material, size and the forces applied.

Deformation in one dimension (1D):

Stress:

It is usually defined as force applied to a material per unit area.Stress is a quantity which defines the magnitude of force that cause deformation.

Strain:

Strain is a measure of how much a material deforms (stretches or compresses) when a force is applied to it. It is calculated by dividing the change in length by the original length of the material:

When pulling force on an object is applied which causes elongation, like the stretching of an elastic band, we call such stress a tensile stress. When forces cause a compression of an object, we call it a compressive stress. As illustraded in Figure 21.2.

The greater the stress, the greater the strain; however, the relation between strain and stress does not need to be linear. When stress is sufficiently low, deformation is causesd in direct proportion to the stress value. The proportionality constant in this relation is calledthe **elastic modulus**. In the linear limit of low stress values, the general relation between stress and strain is known as Hooke's law is given as under

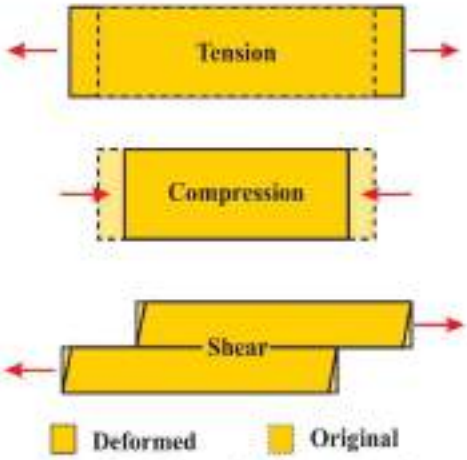


Figure 21.2 Deformation in one dimension

Stress =(elastic modulus)× Strain

We can also see from above equation that when an object is characterized by a large value of elastic modulus, the effect of stress is small. On the other hand, a small elastic modulus means that stress produces large strain and noticeable deformation happens. For example, a stress on a rubber band produces larger strain (deformation) than the same stress on a steel band of the same dimensions because the elastic modulus for rubber is two orders of magnitude smaller than the elastic modulus for steel.

DO YOU KNOW?

The zircons, the oldest known terrestrial materials, have helped portray how the Earth's crust formed during the first geologic eon of the planet. "This may also help us understand how other habitable planets would form.

21.2 Mechanical Properties of Solids

Mechanical properties of solids refer to the characteristics that describe how a solid material responds to applied forces or deformation; some important mechanical properties of solid are Elasticity, Ductility, Brittleness, and Hardness etc.

21.2.1 Young's Modulus (modulus of elasticity)

It is a mechanical property of solids that measures the stiffness or elasticity of a solid material and it computes how much a material will deform (stretch or compress) under a given force.

It is denoted by the symbol Y and has units of $Y = \frac{\text{Force}}{\text{Area}}$ or (Pascal, Pa or N/m^2).

It is a measure of a material's resistance to elastic deformation when subjected to an external force. The higher the Young's Modulus, the stiffer the material.

For example, steel has a higher Young's Modulus than rubber, which makes steel more resistant to deformation than rubber when subjected to a force.

The amount of extension of an object depends upon

i) Applied force
ii) Material
iii) Dimension

If we compare rods made of the same material having different lengths and cross-sectional areas, it is found that the longer the object, the more it elongates for a given stress; and the thicker it is, the less it elongates because strain is proportional to the original length and inversely proportional to the cross-sectional area.

Then Hooke's law can be written for Young's Modulus is defined as the ratio of stress to strain.

$$Y = \frac{Stress}{Strain}$$

$$Y = \frac{\frac{F}{A}}{\frac{\Delta L}{L}}$$

$$Y = \frac{F \times L}{\Delta L \times A} \quad \ldots\ldots\ldots (21.1)$$

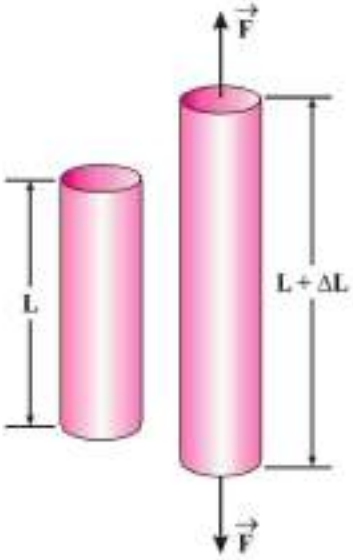


Figure 21.3 longitudinal stress

Worked Example 21.1

Calculate the change in length of the upper leg bone (the femur) when a 70.0 kg man supports 62.0 kg of his mass on it, assuming the bone to be equivalent to a uniform rod that is 40.0 cm long and2.0 cm in radius. Young's modulus for bones is 19×10^{9} N/m^2.

Solution:

Step 1: data

Mass of man = 70.0 kg
Mass supported by leg = 62.0 kg
Length of leg, L = 40.0 cm = 0.4 m
Radius of bone = 2.0 cm = 0.02 m
Young's modulus, Y = 19×10^{9} N/m^2.
Change in length, ΔL = ?

Step 2:

F= W = mg = 62.0 × 9.8 = 607.6 N

Step 3: Calculate the cross-sectional area:

$A = \pi r^2 = (3.14)(0.02)^2 = 12.56 \times 10^{-4} m^2$

Now calculate the change in length

$$Y = \frac{F \times L}{\Delta L \times A}$$

$$\Delta L = \frac{F \ x \ L}{A \ x \ Y}$$

$$= \frac{607.6 \times 0.40}{19 \times 10^{9} \ \times \ 1.256 \times 10^{-4}}$$

$$\Delta L = 1.01 \times 10^{-5} \, m$$

Result: $\Delta L = 1.01 \times 10^{-5} \, m$

Self-Assessment Questions:

1. Calculate the amount of stretch in steel cable. Suspension cables are used to carry gondolas at Murree Hills. Consider a suspension cable that contains an unsupported span of 1.5 km with diameter of 3.0 cm, with maximum tension it can withstand is 3.0×106N.

21.2.2 Shear Modulus:

Shear modulus, also known as the modulus of rigidity is a mechanical property that describes a material's resistance to shear deformation when a force is applied perpendicular to the material's surface. It is denoted by the symbol G or S and has units of force per unit area, typically expressed in Pascal (Pa) or Giga Pascal (GPa).

Shear modulus is the measure of the rigidity of the body which is the ratio of shear stress to shear strain.

Consider a rigid body as shown in figure 21.4. When acted upon by tangential force to twist it.

The shear stress is F/A and shear strain is ΔX/L. Therefore, the shear modulus becomes

$$G = \frac{\frac{F}{A} \times L}{\Delta x}$$

Table: No 21.1

Material	G($10^9 Nm^2$ or GPa)
Aluminum	25
Brass	36
Copper	42
Glass	23
Iron	70
Lead	5.6
Nickel	77
Steel	84
Tungsten	150
Wood	10
Diamond	450
Rubber	0.1

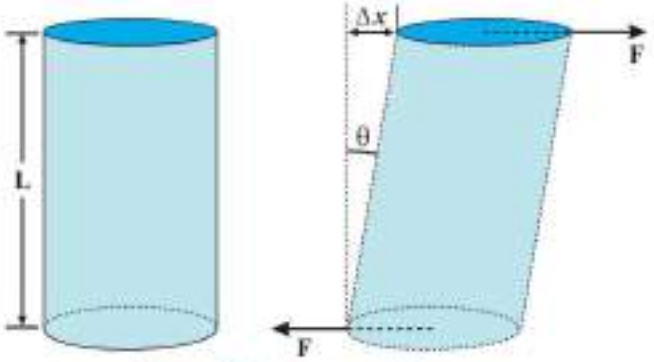


Figure 21.4: shear modulus

$$ShearingStrain = \Delta X = \frac{tan\theta}{l} \ldots\ldots\ldots (21.2)$$

Worked Example 21.

The area of the upper face of a rectangular block is 0.25 m^2 and the lower face is fixed. The height of the block is 1.5 cm, a shearing force applied to the top face which produces a displacement of 0.0152 mm. Find the strain, stress and the shearing force. Modulus of rigidity is 4.5×10^{10} N/m^2.

Solution:

Step 1:data

$Area\ under\ shear\ =\ 0.25\ m^2$

$Height\ of\ block\ =\ 1.5\ cm\ =\ 1.5\ \times10^{-2}\ m$

$Displacement\ of\ top\ face\ =\ 0.0152\ mm\ =\ 0.0152\ \times10^{-3}\ m$

$Modulus\ of\ rigidity\ =G\ =\ 4.5\ \times10^{10}\ N/m^2.$

$Shear\ Strain\ =?$

$Shear\ Stress\ =?$

Shearing Force =?

Step 2: Formula:

$Shearing\ Strain\ =\ \Delta X/L$

Step 3: Calculations: Shearing Strain

$$Shearing\ Strain = \frac{0.0152\times10^{-3}}{1.5\times10^{-2}}$$

Shearing Strain = 1.01×10^{-3}

Shear Stress = G × Shearing Strain

Shear stress= $4.5\times10^{10}\times 1.01\times10^{-3}$

Shearing Stress = 4.56×10^{7} N/m^2

Shear Stress = F /A

F = Shear Stress × A

F = $4.56\times10^{7}\times 0.25$

F = 1.14×10^{7} N

Result: F = 1.14×10^{7} N

21.2.3 Bulk Modulus:

Bulk modulus, also known as the modulus of compressibility, is a mechanical property that describes the resistance of a material to compression under uniform external pressure. It is denoted by the symbol K and has units of pressure (usually expressed in Pascal, Pa or Giga Pascal (GPa).

Bulk modulus is defined as the ratio of the change in pressure applied to a material to the resulting fractional change in volume that occurs in the material. Mathematically, it is expressed as:

$$\frac{\Delta V}{V} = -\frac{\Delta P}{B}$$

DO YOU KNOW?

Compressibility

The reciprocal of bulk modulus of elasticity is called as compressibility

$Compressibility = 1/B$

Its SI unit is m2 /N or Pa.

Material	Bulk modulus
Air	100KPa
Water	2.2 GPa
Steel	150GPa
Diamond	443 GPa

$$B = -\frac{\Delta P}{\frac{\Delta V}{L}} \quad \text{..... (21.3)}$$

where B is the bulk modulus, ΔP is the change in pressure applied to the material, ΔV is the resulting change in volume, and V is the original volume of the material.

DO YOU KNOW?

Viscoelastic materials combine the properties of rubber and slush, deforming and flowing when stressed. They absorb and dissipate energy over time, making them ideal for shock absorbers, vibration reduction, and noise dampening. Examples include rubber, plastics, oil, and molten glass, which are valuable in engineering for designing durable products like gaskets and automotive parts that withstand repeated stress and deformation.

21.2.4 Force– Extension Graph:

In material science, the force versus extension graph for a material gives the relationship between stress and strain. This is obtained by gradually applying a load to a test component and measuring its deformation according to tensile standards. These curves reveal many of properties of materials, such as the young's modulus, the yield strength, the ultimate tensile strength and so on. For different types of materials as shown in figure 21.5,

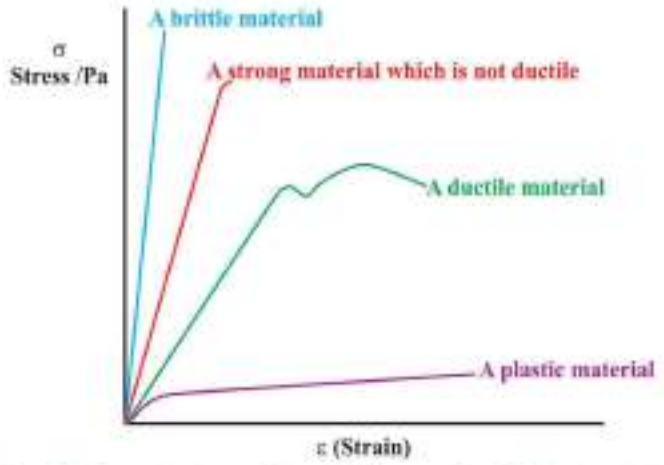


Figure 21.5 force extension graph reveals the properties of different materials

The blue line (brittle) represents the behavior of a typical elastic material, which follows Hooke's Law. As the material is stretched, the force applied to it increases linearly with the extension, up to a point called the yield point. The material eventually reaches a maximum point of extension, is called the ultimate extension or fracture point, beyond which it breaks or fractures.

DO YOU KNOW?

Silk is a natural polymer that is widely used worldwide in textiles. Silk threads are extremely light, but are very strong relative to their weight. It is extremely ductile, so it can bend and stretch without breaking.

The green line (ductile) represents the behavior of a typical viscoelastic material, which exhibits time-dependent deformation under a constant force. In this case, the material initially undergoes elastic deformation, similar to the elastic material, but then continues to deform slowly over time under a constant force, eventually reaching a maximum extension or creep limit.

The purple line (Plastic) represents the behavior of a typical plastic material, which exhibits permanent deformation when subjected to a force. In this case, the material undergoes plastic deformation immediately, with no linear region or yield point, and continues to deform permanently as the force is increased, until it ultimately fractures or breaks.

The Red line is showing strong material which is not ductile, like steel wires stretch very little and breaks suddenly.

Polymeric Materials:

Polymer curves are produced by stretching a sample at a constant rate through the application of a tensile force. By using a constant rate of testing the strain-rate dependency of polymer behavior is not allowed to dominate. However, it should be appreciated that polymers have a marked inherent time - dependence in their response to deformation, which sets their behavior apart from other classes' materials.

21.2.5 Graph of Ultimate tensile stress, elastic deformation and plastic deformation:

Figure 21.6, represents the graph of applied force versus elongation up to a point is called the proportional limit. Hooke's law is the best approximation for many common materials, and the curve is a straight line. Beyond this point, the graph deviates from a straight line, and no simple relationship exists between F and Δl. Nonetheless, up to a point farther along the curve called the elastic limit, the object will return to its original length if the applied force is removed. The region from the origin to the elastic limit is called elastic region. If the object is stretched beyond the elastic limit, it enters the plastic region: it does not return to the original length upon removal of the external force, but remains permanently deformed (such as bent paper clip shown in figure 21.7) such is elastic deformation. The maximum force that can be applied without breaking is called ultimate tensile strength of the material. If the stretching of bonds happens but the atoms do not slip past each other, when the stress is sufficient to permanently deform the metal i.e., soft metal like silver, gold, it is called plastic deformation.

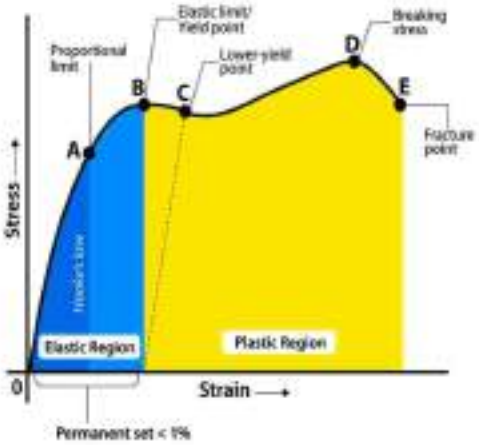


Figure 21.6 Applied force vs elongation

Figure 21.7 elastic deformation

21.3 Electrical Properties of Solids:

The electrical property of a good conductor depends on its ability to transmit energy without boiling, melting, or changing its composition in any way.

- **Resistivity:** The resistance per unit length and cross-sectional area is called resistivity. It varies from material to material. Its SI unit is Ohm-meter (Ω - m).
- **Conductivity:** It is a measure of a material's ability to conduct electric current, which is influenced by the availability and mobility of charge carriers within the solid.
- **Temperature Coefficient of resistivity:** The temperature coefficient of resistance is defined as the change in electrical resistance of a substance with respect to per degree rise in temperature. When the temperature increases, the process of electron collision becomes rapid and faster. As a result, the resistance will increase with the rise in temperature of the conductor.
- **The quantity of charge carriers' n,** the number of chare carriers per unit volume, can be found from measurement of the Hall Effect. Its SI unit is m^{-3}.

21.3.1 Energy Bands in Solids:

When isolated atoms comes together to form a solid, interactions between neighboring atoms cause the electron energy levels to split and overlap, creating continuous energy bands.

Valence Band:

The valence band is the range of energy levels occupied by electrons that are bound to atoms and participate in chemical bonding. It is the highest energy band that contains electrons under normal conditions.

Forbidden Band (Energy Gap):

The forbidden band, or energy gap, is the range of energy levels between the valence band and the conduction band where no electron states exist. This gap influences the material's electrical properties; a larger gap typically means the material is less conductive.

Conduction Band:

The conduction band is the range of energy levels where electrons are free to move throughout the material, allowing electrical conduction. Electrons in the conduction band are not bound to any particular atom and can carry electric current through the solid.

In the case of a diamond crystal, carbon atoms with the electronic structure $1s^2\ 2s^2\ 2p^2$ form two main energy bands as shown in figure, 21.8. The valence band and the conduction band are separated by an energy gap known as the "forbidden band." The valence band contains the lower

DO YOU KNOW?

The Pauli Exclusion principle states that "No two electrons in a given interacting system may have the same quantum state". Thus there must be a splitting of the discrete energy levels of the isolated atoms into new level belonging to the pair, other than to individual atoms

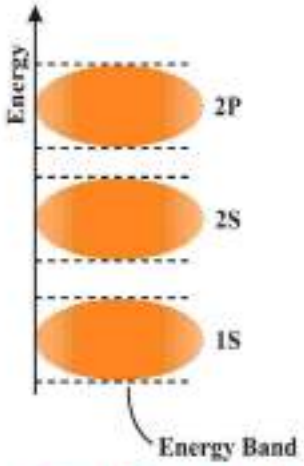


Figure 21.8 energy Bands

energy states, while the conduction band contains higher energy states. This energy gap prevents electrons in the valence band from easily moving to the conduction band, significantly influencing the electrical properties of the solid, such as its ability to conduct electricity.

21.3.2 Classification of Solids:

Depending on the electrical conductivity solids are classified into three main categories: conductors, semiconductors and insulators, as shown in figure 21.9.

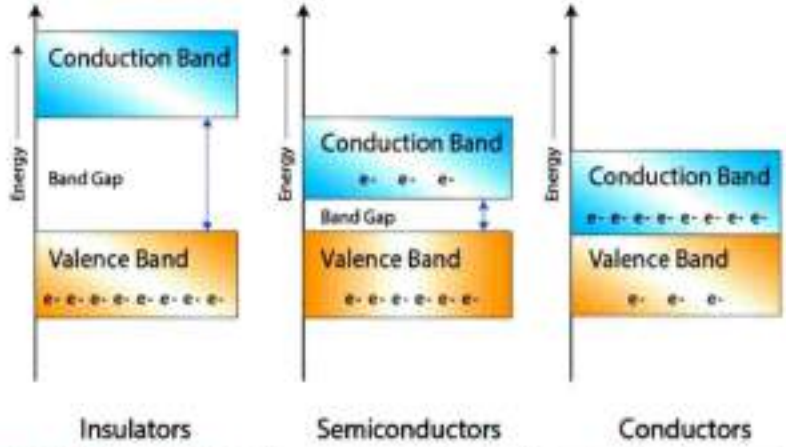


Figure 21.9 Classifications of Solids in three main categories: Insulators, Semiconductors and conductors

Insulators:

These materials do not conduct electricity. The band gap between the valence band and conduction band is very large. Even if a large amount of energy is provided to these solids, they do not conduct electricity. such as wood, plastics etc.

Conductor:

There is no gap between the conduction band and the valence band. Thus, electrons can easily flow from the valence band to the conduction band under the influence of an electric field, making them good conductors of electricity.

Semiconductors:

The gap between the conduction band and the valence band is very less; therefore, whenever sufficient energy is provided to the electrons in a semiconductor, electrons jump from the valence band to the conduction band. The conductivity of the semiconductors increases with an increase in temperature, for semiconductors, it lies in between 10^{-6} to 10^{-4} $(\Omega\ m)^{-1}$. There are two types of semiconductors:

a) Intrinsic semiconductor:

When energy is provided to intrinsic semiconductors like silicon and germanium, electrons leave their positions,

DO YOU KNOW?

Measurements show that the resistivity of super conductors is less than 4×10^{-25} Ω m, which is over 1016 times smaller than copper, and is considered to be zero in practice.

creating positive holes. Under an electric field, electrons and holes move in opposite directions, enabling electrical conductivity. Group IV elements are considered as an interinsic semiconductors

b) Extrinsic semiconductor:

Silicon and germanium have low conductivity at room temperature. Adding small amounts of group III or group V impurities, a process called doping, increases their conductivity, resulting in extrinsic semiconductors.

Intrinsic and extrinsic semiconductor are shown in Figure 21.10

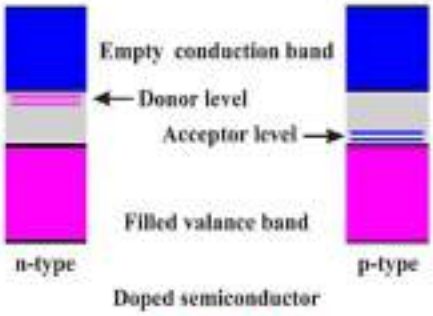


Figure 21.10
Intrinsic and extrinsic semi-conductor

21.3.3 Superconductor:

In 1911, Dutch physicist Heike Kamerlingh Onnes discovered that mercury's resistivity disappears below about 4 Kelvin.

A superconductor is a material that can conduct electricity without resistance when it is cooled below a certain critical temperature, typically near absolute zero. As shown in figure 21.11

This allows charge to flow without energy loss to thermal effects. Superconductors exhibit this behavior below a critical temperature, usually near absolute zero, with resistivity less than $4 \times 10^{-25} \Omega$-m. They also exhibit perfect diamagnetism, expelling magnetic fields in the superconducting state, known as the Meissner effect, and enabling phenomena like magnetic levitation.

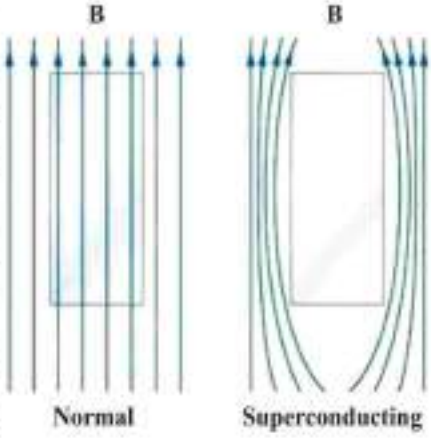


Figure 21.11 supper conductor graph

21.3.4 Super conductor Applications:

Superconductors are used in particle accelerators, generators, transportation, computing, electric motors, medical applications, and power transmission, among other fields.

1. Magnetic Resonance Imaging (MRI):

MRI scanners provide high resolution picture of the tissues inside the body. Superconducting coils produce a strong magnetic field (up to 60,000 times as strong as the intensity of Earth's magnetic field) that is used to align the protons of hydrogen atoms in the body of the

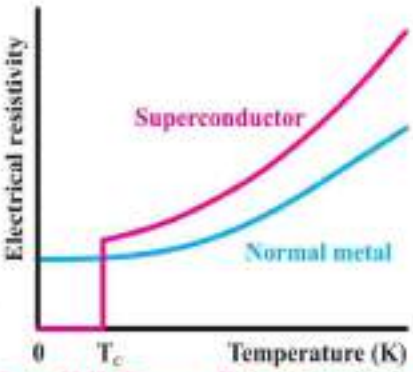


Figure 21.12 Meissner effect super conductors

patient, as shown in figure 21.13. Like electrons, protons have a 'spin' property, so they align with a magnetic field the proton's axis vibrates about the applied magnetic field. Vibrating protons are crashed with a burst of radio waves tuned to push the Protons' spin axes are sideways, perpendicular to the applied magnetic field. When radio waves pass and the protons quickly return to their vibrating pattern, they emit a faint electromagnetic signal whose frequencies depend slightly on the chemical environment in which the proton resides. These signals, detected by sensors, are then analyzed by a computer to reveal varying densities of hydrogen atoms in the body and their interaction s with surrounding tissue. The resulting images clearly distinguish between fluid and bone. MRI was formerly called NMRI (Nuclear Magnetic Resonance Imaging) because hydrogen nuclei resonate with the applied fields.

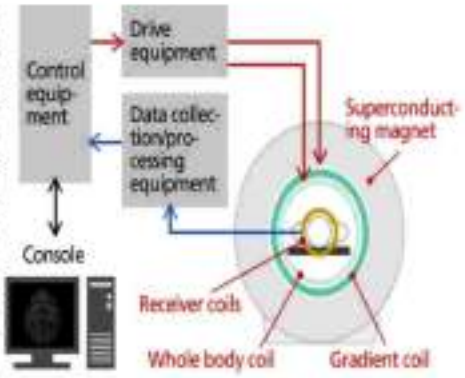


Figure 21.13 working principle of MRI

2. Maglev Train Systems:

Maglev train systems use powerful electromagnets to float the trains over a guide-way, instead of the old steel wheel and track system as shown in figure 21.14. A system called electromagnetic suspension suspends, guides, and propels the trains. A large number of magnets provide controlled tension for lift and propulsion along a track. Maglev (derived from magnetic levitation) is aim of train transportation that uses two sets of magnets: one set to repel and push the train up off the track and another set to move the elevated train ahead, taking advantage of the lack of friction.

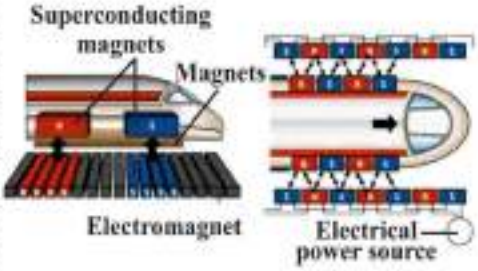


Figure 21.14 Magnetic Levitation

3. Small Electric Motor:

Precision Instruments:

Superconducting motors can be used in precision instruments where high efficiency and low noise are critical.

Medical Devices:

Compact, efficient motors for medical devices such as MRI machines, which already use superconducting magnets.

Aerospace:

High-efficiency, lightweight motors for aerospace applications, where every gram of weight saved is crucial.

Robotics:

Small, powerful motors for advanced robotic applications requiring precise control and high power output.

4. Super Conductivity and Super Computers:

Super conducting components can improve the performance of super computers in several ways as:

Super conducting qubits:

These are the basic building blocks of quantum computers, which are able to perform calculations at speeds that are exponentially faster than classical computers. Superconducting qubits can maintain their quantum states for long periods of time, allowing for complex computations to be performed.

Faster processing speeds:

Super conducting components can process information at much faster speeds than traditional electronic components, resulting in faster computing times and the ability to handle larger datasets.

Lower power consumption:

Super conducting circuits require less power to operate than traditional electronic circuits, reducing energy costs and making them more environmentally friendly.

Improved reliability:

Super conducting components are less prone to errors and can operate for longer periods of time without failure, resulting in more reliable and efficient supercomputers.

However, the use of superconducting components in super computers is still in the experimental stage and faces several challenges.

21.4.1 Magnetic Properties of Solids:

The magnetic properties of solids depend on how their electrons behave in response to an external magnetic field. There are three main types of magnetic materials: ferromagnetic, paramagnetic, and diamagnetic.

Material	Behavior	Examples
Ferromagnetic	Have magnetic moments (tiny magnetic domains) that tend to align parallel to each other. When exposed to an external magnetic field, these materials can become strongly magnetized.	Iron, Cobalt, Nickel.
Paramagnetic	Have individual magnetic moments that are randomly oriented, and they become weakly magnetized when exposed to an external magnetic field	Aluminum Platinum, Titanium
Diamagnetic	Have no intrinsic magnetic moment. When exposed to an external magnetic field, they develop a weak, negative magnetization that opposes the applied field.	Copper, Bismuth.

21.4.2 Magnetic Domain:

The magnetic domain theory was proposed by Weiss in 1907. According to this theory, ferromagnetic material contains a large number of tiny regions, and each region (magnetic domain) consists of atomic magnetic moments that are aligned in a specific direction. Initially, these magnetic domains are randomly oriented, resulting in a weak magnetization where the magnetic moments cancel each other out in the absence of an external magnetic field.

In the presence of a strong external magnetic field, these magnetic domains align in the direction of the field, resulting in strong magnetization. This alignment occurs because the magnetic moments within each domain are parallel due to the influence of the external magnetic field. Paramagnetic, Ferromagnetic and Anti ferromagnetic as shown in Figure 21.15.

(a) Paramagnetic (Randomly oriented order,)
(b) Ferromagnetic (Perfect order)
(c) Anti ferromagnetic (Perfect anti order)

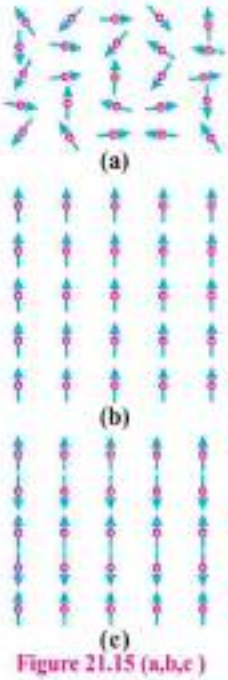


Figure 21.15 (a,b,c) magnetic domain

In ferromagnetic materials, once all the magnetic domains are aligned in one direction and the external field is removed, the domain boundaries do not change their orientation. This property prevents demagnetization, making ferromagnetic materials useful for making permanent magnets. Ferromagnetic materials are strongly attracted to external magnetic fields due to their aligned magnetic domains.

21.4.3 Curie Point:

Curie point is also known as Curie temperature. Some magnetic materials lose their magnetic properties as the temperature rises above the Curie temperature. The Curie temperature weakens the magnetic properties of the material. Pierre Curie, a French physicist, discovered it in 1895 and proposed certain laws related to the magnetic properties of materials as temperature changes.

DO YOU KNOW?

Alloys of iron differ; soft iron is easier to magnetize than steel. It helps to tap the material to nudge any stubborn domains into alignment. Another way is to stroke the material with a magnet. The stroking motion aligns domains. If a permanent magnet is dropped or heated outside the strong magnetic field from which it is made, some of the domains are jostled out of alignment and the magnet becomes weaker.

Curie's Law:

The Curie law states that in a paramagnetic material, the material's ***"magnetization is directly proportional to an applied magnetic field". $M \propto H$***

$M = \chi H$ ……. (H is the applied magnetic field strength)

But when it is heated, the relation is reversed, i.e., the magnetization becomes inversely proportional to temperature.

Mathematically, it is written as:

$$M = \frac{C \times H}{T} \text{.....(21.4)}$$

Where:
M is the magnetization; H is the applied magnetic field, T is the absolute temperature, and C is the Curie constant specific to the material.
The various experiments by Pierre Curie showed that for many substances, the susceptibility χ (chi) is inversely proportional to the absolute temperature T. $\chi = \frac{C}{T}$ .....(21.5)

Consider an example of iron atoms with a temperature of 770°C (1043 K). Each iron atom acts like a tiny magnet at this temperature spontaneously. Each of them will align themselves as some kind of magnetic material.

For pure iron, as shown in Figure 21.16, the atomic magnets are distributed within each microscopic domain. Pure iron is a kind of ferromagnetic material. The directions of the magnetic fields are the same so that their magnetic fields strengthen each other.

Anti-ferromagnetic materials possess atomic magnets that have alternate properties of magnetic fields. They act in opposite directions so that their magnetic fields cancel each other.

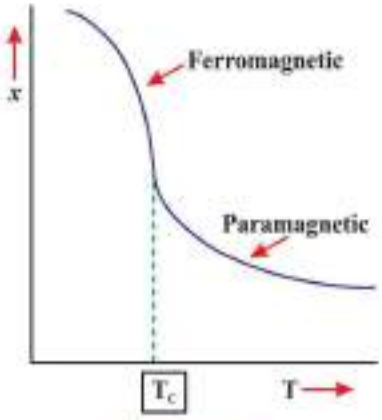


Figure 21.16
critical temperature and curies law

The graph depicts the rise of temperature above the Curie point T_c, which can lead to the production of roughly similar patterns of decreasing magnetic property. This behavior is constant in all three classes of materials.

21.4.4 Hard and Soft Ferromagnetic Substances:

Ferromagnetic materials fall into two main categories: Hard and Soft.

Hard ferromagnetic:

Materials like neodymium, samarium cobalt, and alnico are permanent magnets with high coercivity, suitable for stable magnetic fields in applications like speakers and generators.

Soft ferromagnetic:

Including iron, nickel, and cobalt, these materials have low coercivity, making them easy to magnetize and demagnetize. They are used in devices like transformers and inductors that require rapidly changing magnetic fields. Microstructure influences magnetic properties, with hard substances having a fine-grained structure and soft ones having a coarse-grained structure. The choice between hard and soft ferromagnetic materials depends on the specific application's requirements and operating conditions. For instance, hard ferromagnetic materials are used for permanent magnets in motors, while soft ferromagnetic materials are employed in rotating armatures experiencing rapidly changing magnetic fields.

Table of Curie temperature

TABLE No:21.2
Critical temperature of different elements

Material	Curie T_c(K)
Fe	1043
Co	1388
Ni	627
Gd	293

21.4.5 Hysteresis Loss:

When a material is not magnetized, the axes of different domains point in various directions, resulting in a zero magnetic effect. Upon the application of an external magnetic force, the axes align with the force, creating a strong magnetic field. Repeated cycles of magnetization and demagnetization disturb domain alignment, leading to hysteresis loss. Energy is stored during field establishment and returned during collapse, but hysteresis causes incomplete energy recovery, resulting in heat. This adverse loss, seen in transformers, reduces overall efficiency. To minimize hysteresis loss, soft iron cores are used due to their lower energy loss. Factors affecting hysteresis loss include the area under the hysteresis loop, frequency, and the volume of the material. The loss is directly proportional to these factors, impacting the overall efficiency of devices like transformers.

21.4.6 Synthesis of hysteresis Loop between Magnetic Field Strength and Magnetizing Current.

The hysteresis loop is a graphical representation, as shown in Figure 21.17 that shows the relationship between the magnetic field strength (B) and the magnetization (H) of a material. The loop is particularly useful for understanding how a material responds to changes in an external magnetic field. Here's how you can synthesize the variation of magnetic field strength with magnetizing current from the hysteresis loop.

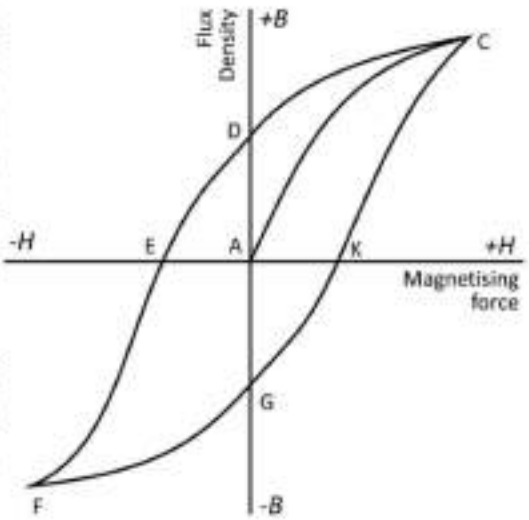


Figure 21.17
hysteresis loop for magnetic material

1. Hysteresis Loop:

A hysteresis loop is a closed curve illustrating a material's magnetic behavior under a changing magnetic field, formed by plotting the material's magnetization against the magnetic field strength.

2. Variation of Magnetic Field Strength with Magnetizing Current:

The horizontal axis of the hysteresis loop represents the magnetizing current or applied magnetic field strength. The vertical axis represents the resulting magnetization of the material.

3. Magnetizing and Demagnetizing:

When the magnetizing current increases, the material gets magnetized, and the magnetic field strength goes up, shown on the right side of the hysteresis loop.

When the magnetizing current decreases, the material may not fully demagnetize, and the magnetic field strength remains at a certain level, shown on the descending left side of the loop.

4. Understanding the Loop:

The width of the hysteresis loop indicates energy loss (hysteresis loss) during magnetization and demagnetization cycles.

The shape of the loop provides information about the material's magnetic properties, such as coercivity and remanence.

By analyzing the hysteresis loop, the relationship between magnetic field strength and magnetizing current can be observed, aiding in understanding the material's magnetic properties and response to external magnetic field changes.

21.4.7 Hysteresis loop for permanent and temporary magnets:

Hysteresis is especially marked in materials of high residual magnetism, such as hardened steel. In most cases, hysteresis is a necessity as it causes dissipation of heat, waste of energy, and vibration due to change in polarity and rotation of element magnets in the material.

Hysteresis loops for hard steel, wrought iron, cast steel, and for alloyed sheet steel are shown in Figure 21.18.

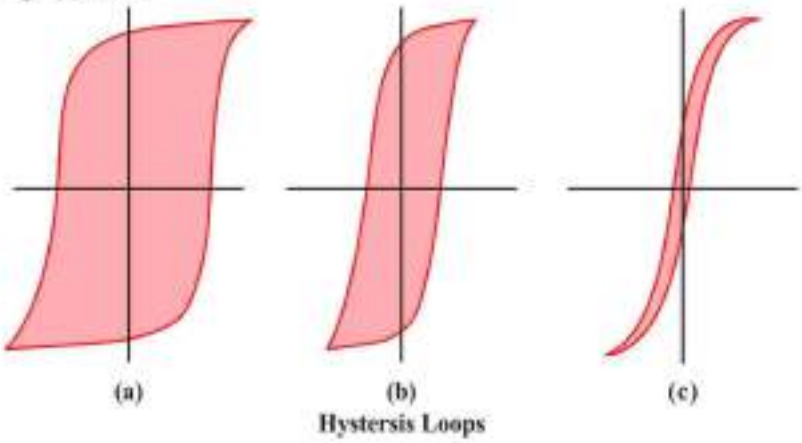


Figure 21.18(a, b, c)
hysteresis loops for hard steel, wrought iron and cast steel and low carbon steel respectively

Loop(a) is for hard steel. Due to its high retentivity power and large coercive force, this material is well-suited for permanent magnets. Since the area of the hysteresis loop for hard steel is large, hard steel is not suitable for rapid reversals of magnetization. As shown in Figure 21.18(a)

Loop (b) is for wrought iron and cast steel, which rise steeply. Hence, these materials have high magnetic permeability and good retentivity; therefore, these materials are suitable for cores of electromagnets. As illustrated in Figure 21.18(b)

Loop(c) is for iron, low carbon steel, silicon alloys, perm-alloy (alloy of Nickel and Iron), or Mu-metal (Nickel-Iron soft ferromagnetic material) sheets. Since the permeability of these materials is very high and hysteresis losses are very low, these materials are most suitable for transformer cores and armatures, which are subjected to rapid reversals of magnetization. As shown in Figure 21.18(c).

SUMMARY

- A crystalline solid is one where the internal structure is regular in nature, the atoms within it being set in well-defined patterns.
- Amorphous materials do not possess repeated structures over a long range as in crystalline solids. Rubber is an example of a polymer. It consists of very long chain molecules.
- Stress is a quantity which defines the magnitude of forces that cause deformation. Stress is usually defined as force per unit area.
- When forces pull on an object and cause its elongation, like the stretching of an elastic band, we call such stress a tensile stress.
- Strain is given as a fractional change in either length (under tensile stress) or volume (under bulk stress) or geometry (under shear stress).
- The tensile strength of a solid is a measure of its resistance to being pulled apart.
- Ductility is a property of metal that enables it to be drawn through a die to produce a wire.
- Elasticity is a measure of a deformed object's ability to return to its original size and shape once the outside forces are removed.
- Young's modulus is defined as the ratio of stress to strain.
- Bulk Modulus If an object is subjected to inward forces from all sides, its volume will decrease.
- Energy Band: The electrons in the same orbit exhibit different energy levels. The grouping of these different energy levels is known as an energy band.
- Superconductivity allows charge to flow through a superconducting conductor without losing its energy to thermal energy.
- Magnetic Hysteresis / Retentivity: When a magnetic material is placed in an external magnetic field due to which its grains get oriented in its direction and the material is magnetized. If this material is removed from that external magnetic field, even some magnetization remains in that material which is called residual magnetism. This is also called retentivity of the material.
- A magnetic domain is a region within a magnetic material in which magnetization is in a uniform direction.
- Curie point or Temperature is defined as when some magnetic materials lose their magnetic property as temperature becomes above the Curie temperature.
- Hysteresis Loss is the disturbance in the alignment of the various domains causes hysteresis loss.

3. If the load is increased beyond the point, the strain increases rapidly for even a small change in the stress.
   (a) Elastic point (b) Yield point (c) Plastic point (d) Fracture point
4. The reciprocal of the bulk modulus is called
   (a) Compressibility (b) Volume stress
   (c) Modulus of rigidity (d) Volume strain
5. Which of the following statements is/are wrong?
   i. Hollow shaft is much stronger than a solid of the same length and same mass.
   ii. Reciprocal of bulk modulus of elasticity is called compressibility.
   iii. It is difficult to twist along rod as compared to small rod.
   (a) III only (b) I only (c) II and III (d) I and II
6. Metals are good conductors of heat and electricity. This property is conferred by
   (a) Covalent (b) Ionic (c) Metallic (d) Hydrogen
7. For a metallic crystal, delocalized electrons occupied band is:
   (a) Conduction band (b) Valence band
   (c) Conduction and valence bands (d) there are no delocalized electrons
8. The semiconductors have resistivity
   (a) Between conductor and insulator (b) More than insulators
   (c) Less than conductors
   (d) Depending upon the semiconductor material property.
9. A ferromagnetic substance becomes a permanent magnet when it is placed in a magnetic field because
   (a) All the domains get oriented in the direction of the magnetic field.
   (b) All the domains get oriented in the direction opposite to the direction of the magnetic field.
   (c) Domains get oriented randomly.
   (d) Domains are not affected by the magnetic field.
10. Identify methods to demagnetize a ferromagnet.
   (a) By cooling, heating, or submerging in water
   (b) By heating, hammering, and spinning it in an external magnetic field
   (c) By hammering, heating, and rubbing with cloth
   (d) By cooling, submerging in water, or rubbing with cloth

Section (B): CRQs (Short Answered Questions):

1. Why are the springs made of steel and not of copper?
2. The breaking force for a wire is F. What will be the breaking force for two parallel wires of the same size?
3. Distinguish between intrinsic and extrinsic semiconductor.
4. A wire is replaced by another wire of same length and material but of twice the diameter. What will be the effect on the:
   (a) Increase in its length under a given load? (b) Maximum load which it can bear?
5. Sand does not possess any definite shape and volume, still it is solid. Give reason.
6. Specify the importance of stress-strain curve.
7. Why liquids don't possess rigidity?

8. Give applications of Curie point.
9. What are amorphous materials and what are their uses?

Section (C): ERQs (Long Answered Questions):

1. Explain Force-Extension graph.
2. Derive relation for Young's Modulus and Shear Modulus.
3. Distinguish between structure of crystalline, glassy, amorphous, and polymeric solids.
4. Describe the energy bands in solids.
5. Describe superconductivity and its applications.
6. Discuss the applications of superconductors for MRI, Maglev's, and supercomputers.
7. Describe hysteresis loss.
8. Synthesize hysteresis loop for relationship between magnetic field strength and magnetizing current.
9. Discuss energy bands and their classification. Explain magnetic properties of soft and hard magnetic materials.

Section (D): Numerical:

1. The 'lead' in pencils is a graphite composition with a Young's modulus of 1.0×10^9 N/m^2. Calculate the change in length of the lead in an automatic pencil if you tap it straight into the pencil with a force of 4.0 N. The lead is 0.50 mm in diameter and 60 mm long. **(1.0mm)**
2. A wire of 2.2 m long and 2.25 mm in diameter, when stretched by a weight of 8.8 kg, its length has been increased by 0.25 mm. Find the stress, strain, and Young's modulus of the material of the wire. Given g = 9.8 m/s^2. **(2.2×10^7 N/m^2, 1.14×10^{-4}, 2×10^{11} N/m^2)**
3. A farmer making juice fills a glass bottle to the brim and caps it tightly. The juice expands more than the glass when it warms up, in such a way that the volume increases by 0.2% (i.e., $\frac{\Delta V}{V_0} = 2 \times 10^{-3}$) relative to the space available. Calculate the normal force exerted by the juice per square centimeter, if its bulk modulus is 1.8×10^9 N/m^2. Assuming that the bottle does not break. **(432 N/cm^2)**
4. The elastic limit of copper is 1.5×10^8 N/m^2. It is to be stretched by a load of 10 kg. Find the diameter of the wire if the elastic limit is not to be exceeded. **(0.912 mm)**
5. What would be the greatest length of a steel wire which is fixed at one end, and can it be hanged freely without breaking? The breaking stress of steel is 7.8×10^8 N/m^2, and the density of steel is 7800 kg/m^3. **(1.02×10^4 m)**
6. A mild steel wire of radius 0.55 mm and length 3.5 m is stretched by a force of 52 N. Calculate: (a) Longitudinal stress, (b) Longitudinal strain, and (c) Elongation produced in the wire if Young's modulus is 2.1×10^{11} N/m^2.
   (5.47×10^7 N/m^2, 2.6×10^{-4}, 0.91 mm)
7. Calculate the change in volume of a lead block of volume 1.3 m^3 subjected to a pressure of 12 atm. Also, calculate the compressibility of lead. Given the bulk modulus as $B=80\times10^9$ N/m^2. **(1.97×10^{-5} m^3, 1.25×10^{-11} N/m^2)**
8. The thickness of a metal plate is 0.35 inches. It's drilled to have a hole of radius 0.08 inches on the plate. If the shear strength is 4×10^4 lbs/in^2, determine the force needed to make that hole. **(0.176 in^2, 7×10^3lbs)**

Introduction:

Solid-state electronics is a branch of electronics that deals with the flow of electric current through solid materials, such as semiconductors, insulators, and conductors. It involves the design, development, and application of electronic devices and circuits that use these materials to control and manipulate electrical energy. Solid-state electronics has revolutionized the field of electronics, enabling the creation of smaller, faster, and more efficient devices, such as transistors, diodes, integrated circuits, and microprocessors, which are essential components in modern computers, smart phones, and other electronic devices. The field is constantly evolving, with advancements in materials science, nanotechnology, and quantum mechanics.

22.1.1 Intrinsic (Pure) and Extrinsic (Doped) Semiconductors:

Intrinsic (Pure) Semiconductors:

Intrinsic semiconductors are materials that are chemically pure, meaning they do not contain any significant amount of impurity atoms. The two most commonly used intrinsic semiconductors are silicon (Si) and germanium (Ge). These materials have a crystalline structure, and their electrical properties are determined by the arrangement and behavior of their atoms.

Extrinsic (doped) Semiconductors:

Doped semiconductors are created by introducing impurity atoms into the intrinsic semiconductor material. This process is known as doping, and it significantly alters the electrical properties of the semiconductor. Doping can be done with two types of impurities, resulting in two types of doped semiconductors: n-type and p-type.

In summary, intrinsic semiconductors are pure materials with moderate conductivity, whereas doped semiconductors are intentionally infused with impurities to enhance their electrical properties, making them essential for modern electronic devices.

22.1.2 Production of P-Type and N -Type Semiconductors:

The intentional addition of impurities alters the conductivity and electrical properties of the semiconductor, allowing for the controlled operation of charge carriers.

N-type Semiconductors:

It is characterized by an excess of positively charged carriers, known as "holes," or majority Positive charge carriers.

In *P*-type semiconductors, the holes effectively behave as positive charge carriers in the material. These holes are essentially vacancies in the crystal lattice structure, where an electron is missing from a **covalent bond**.

Semiconductor material, silicon (Si) or germanium (Ge) is the base semiconductor material.

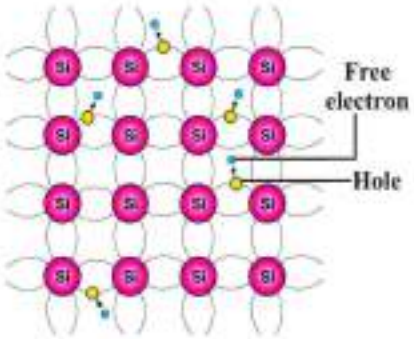


Figure 22.1
free electrons and hole in semiconductor

Doping with Donor Impurities:

A small amount of a group-V element (penta-valent) is doped into the semiconductor's crystal lattice, as illustrated in figure 22.2. Common donor impurities include phosphorus (P) or arsenic (As).

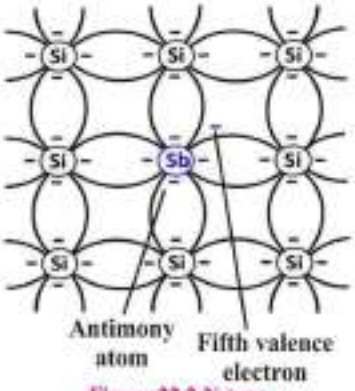


Figure 22.2 N-type

DO YOU KNOW?

Acceptor Ionization: At normal operating temperatures, thermal energy is sufficient to ionize the acceptor atoms, creating mobile holes in the valence band. These holes contribute to the overall conductivity of the material.

These impurity atoms have one more valence electron than the semiconductor material, creating extra electrons that are free to move and conduct electricity.

P-type Semiconductors:

It is characterized by an excess of negatively charged carriers, namely electrons, or majority negative charge carriers.

In *N*-type semiconductors, electrons are the primary charge carriers responsible for the material's electrical conductivity.

Doping with Acceptor Impurities:

If element of group-III (tri-valent) is added into the crystal lattice of the semiconductor as shown in figure 22.3. Common acceptor impurities include boron (B) or gallium (Ga). These impurity atoms have one fewer valence electron than the semiconductor material, creating "holes" in the crystal lattice where electrons can move.

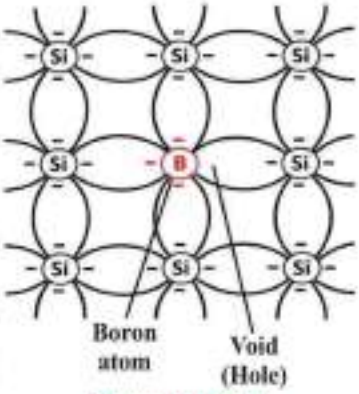


Figure 22.3 P-type

22.1.3 Electron and Hole in Semiconductor:

When an electric field is applied then electric current flow is performed in two ways.

When the temperature is increased then some ***valence electrons*** **enter into the conduction band** and become free electrons. Whenever an electron enters into the conduction band, then a vacant space or a hole is produced in the valence band as shown in figure 22.4. Due to the influence

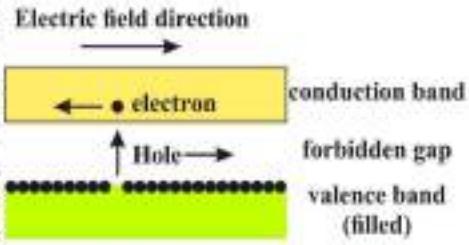


Figure 22.4 electrons and Holes

of potential difference or electric field, flow of electrons and holes are produced.

At room temperature, some electrons from the valence band can cross this small energy gap and are accumulated in the conduction band. As a result, an equal number of holes and electrons are produced. This phenomenon is called electron-hole pair Production.

At the temperature above absolute zero, if the electric field is applied then conduction electrons move towards anode and holes move towards the cathode. So, it can be said that the current flow in the semiconductor is the flow of electrons and holes in the valence and conduction bands in opposite directions.

DO YOU KNOW?

Most familiar intrinsic semiconductors are germanium and silicon. The energy gap between their valence band and conduction band are 0.72 eV and 1.1 eV respectively.

22.2 P-N Junction:

The ***P-N*** junction is formed when the *p*-type and *n*-type semiconductors are joined, is called as *P-N* junction as shown in figure 22.5. *P-N* junction is formed within a single crystal of material rather than just simply joining or fusing together two separate pieces. This junction is a crucial component in electronics and forms the basis for many semiconductor devices like diodes and transistors.

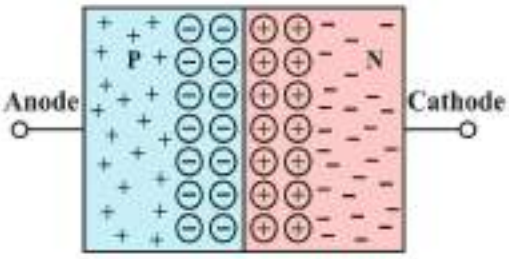


Figure 22.5PN JUNCTION

22.2.1 Flow of Electrons and Holes across Junction:

Initially, a large density rise exists across the junction who leads to a migration of electrons and holes. When electrons move from the N-type to the P-type, negative ions are left behind. Simultaneously, holes from P-type migrate toward the N-type, creating positive ions. This process which is called diffusion, continues until an equilibrium is reached, resulting in a potential barrier zone around the junction, as shown in figure 22.6.

The potential barrier, formed due to repulsion between donor and acceptor atoms, creates a depletion layer. This layer is devoid of free charge carriers, preventing further movement across the junction.

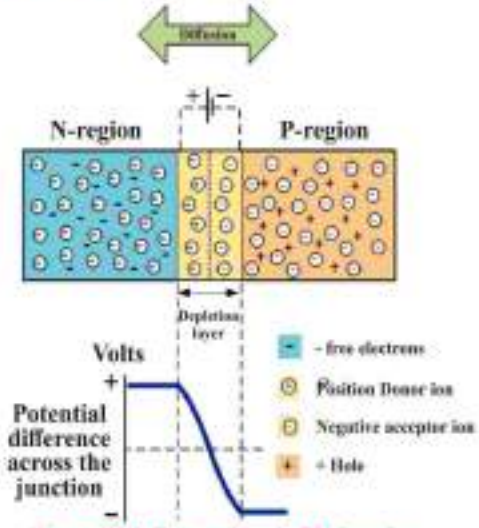


Figure 22.6 Charges flow in PN Junction

22.2.2 P-N Junction as Diode:

A diode is a semiconductor device that is formed through P-N junction and used in allowing the flow of electric current in one direction and blocking in the opposite. The symbol of the *P-N* junction diode is shown in figure 22.7

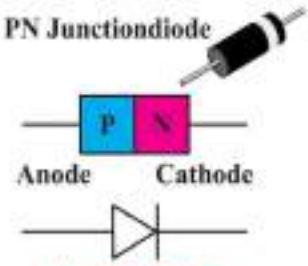


Figure 22.7 Diode

Properties of Diode:

Below are some of the common properties of a diode:

1. Diode has the ability to rectify electric current.
2. It can create a potential barrier and make use of its capacitance properties.
3. Diode creates various nonlinear current-voltage characteristics.

22.2.3 Diode Biasing:

Forward Bias:

When the P-type is connected to the positive terminal of the battery and the N-type is connected to the negative terminal is called Forward bias as shown in figure 22.8.

In this condition, the applied electric field and the built-in electric field at the *P-N* junction are in opposing directions. Adding both the electric fields gives a resultant electric field, which is found to be smaller than the built-in electric field. So the depletion region becomes thinner and less resistant.

When the applied voltage is high, the resistance of the depletion region becomes insignificant. At 0.3V to 0.6 V, the resistance of the depletion region in silicon becomes absolutely insignificant, allowing current to flow freely through it.

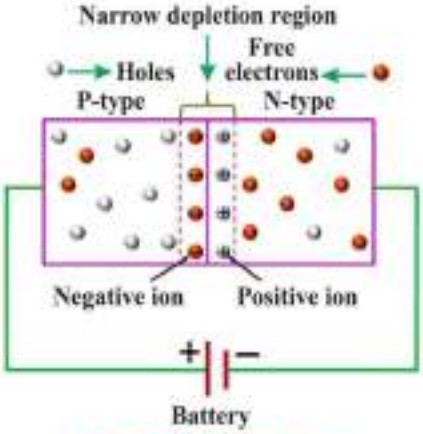


Figure 22.8 Forward Bias

Reverse Bias:

When the P-type is connected to the negative terminal of the battery and the N-type is connected to the positive side is called Reverse Bias as shown in figure 22.9

In this condition, the applied electric field and the built-in electric field are both in the same direction. The resultant electric field and the built-in electric field are also in the same direction, resulting in a more resistive, thicker depletion region. If applied voltage increased, it results thicker and more resistant depletion region.

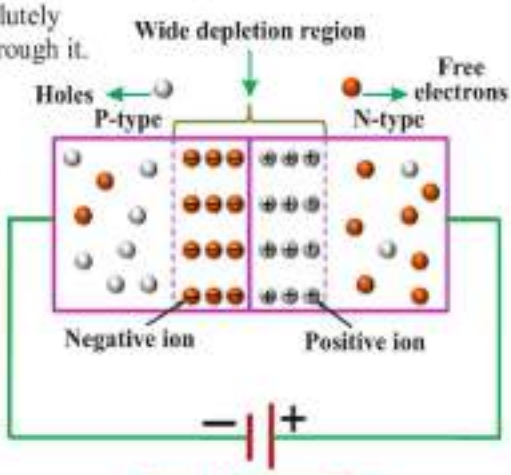


Figure 22.9 Reverse Bias

Self-Assessment Questions:

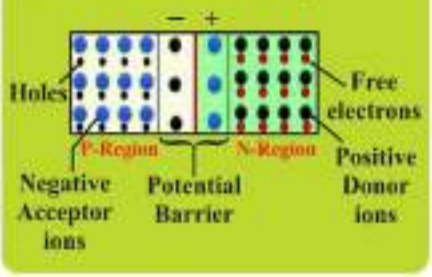
1. What causes the formation of the depletion region in a PN junction at equilibrium?
2. How does forward biasing affect the depletion region and current flow in a PN junction?
3. Why is the current very small in a PN junction under reverse bias?

DO YOU KNOW?

The depletion layer or region acts as a barrier that controls the flow of current. When a voltage is applied, the depletion layer widens or narrows, allowing more or fewer charge carriers to pass through. This mechanism enables the transistor to amplify or switch electronic signals.

The depletion layer is characterized by:

1. Reduced carrier concentration
2. Increased resistance
3. Decreased conductivity



22.2.4 I-V Characteristics of *p-n* Junction:

The relationship between the voltage across the junction and current through the circuit is known as the (V-I) characteristics of a *P-N* junction or semiconductor diode.

The V-I characteristics of the *P-N* junction can be explained in three cases:

- Zero bias or unbiased
- Forward bias
- Reverse bias

At zero bias, no electric current flows through the diode because there's no external voltage applied to enable the movement of electrons or holes. In forward bias as shown in figure 22.10 (a) when the diode voltage (V_d) reaches 0.7 V for silicon and 0.3 V for germanium, current starts flowing. The current increases gradually at first, creating a non-linear curve until the diode surpasses the potential barrier, after which it operates normally and the curve steepens linearly with increasing external voltage.

In reverse bias, only a small leakage current flows, represented to the left of the origin in the graph as shown in figure 22.10(c). This current remains low until the diode breaks down, at which point it can be destroyed unless a high series resistance limits the current.

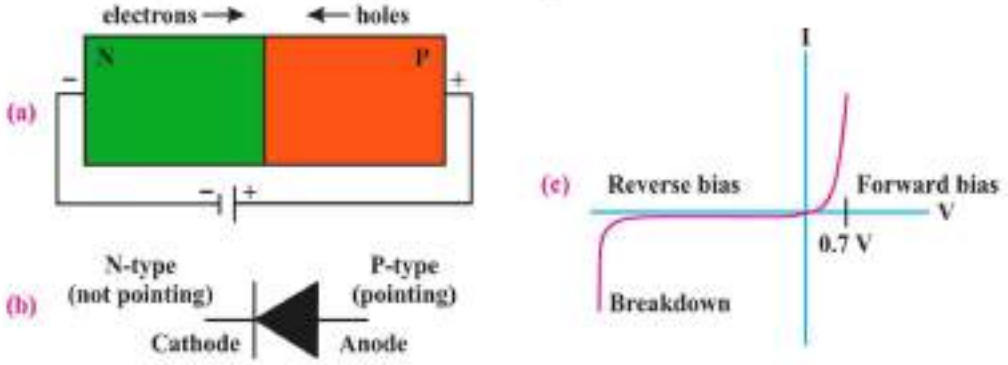


Figure 22.10 I-V Characteristics of *p-n* Junction (a,b,c)

Self-Assessment Questions:

1. Why does reverse saturation current occur in a PN junction?
2. How do the IV characteristics of a PN junction change in forward bias?

22.2.5 Rectification:

Rectification is the process of converting an alternating current (AC) waveform into a direct current (DC) waveform, i.e., creating a new waveform that has only a single polarity as shown in 22.11.

Figure 22.11 Rectification

A diode allows electric current to flow in forward bias condition and blocks the current in reverse bias condition.

Rectification is classified into two types according to the output characteristics which are:

(i) half-wave rectification and
(ii) full-wave rectification.

Half Wave Rectification:

Since a diode allows ***AC*** current to flow only in one direction, it can serve as a rectifier. As shown in figure 22.12(a). The *AC* source applies a voltage across the diode alternately positive and negative. When the positive cycle of AC voltage passes through the diode, the diode is forward biased and act as a closed circuit there is current through the resistor R.

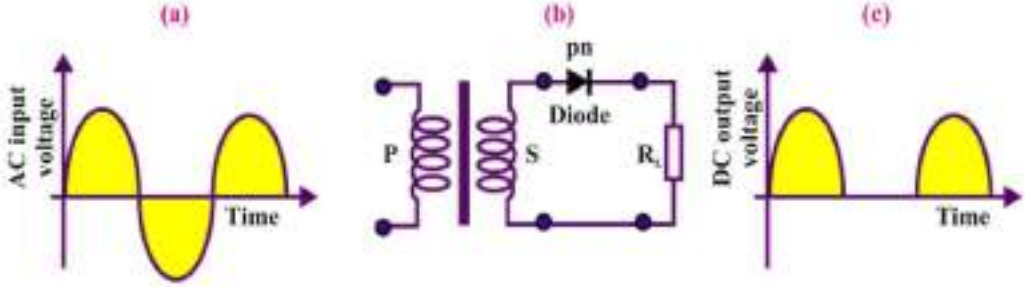


Figure 22.12 (a,b,c) Half Wave Rectifications

During the negative half cycle, the diode is connected with the negative supply which reverses biases the diode, the diode behaves like an open circuit and does not produce the output across the load.

Hence a graph of the voltage V_{ab} across R as a function of time looks likes the output voltage shown in figure 22.12(b). This is called half wave rectification which seems unidirectional from its output signal.

Applications of a Half-wave Rectifier:

- Low power simple battery charger circuit.
- Fire Alarm circuits.
- Soldering Iron circuit.
- Amplitude Modulation (AM) Radio circuits as a Detector.

Full-Wave Rectification:

Full-wave rectification uses two diodes (as shown in figure 22.13(a,b,c).During the positive half cycle of input AC signal, this makes the diode D_1 forward biased (acts as closed switch) and diode D_2 reverse biased (acts as open switch). Therefore, current flows through the load resistor R.

During the negative half cycle of input AC signal, this makes the diode D_2 forward biased and the diode D_1 reverse biased. Therefore, the current will flow through diode D_2 through, load resistor R and lower half of the secondary winding.

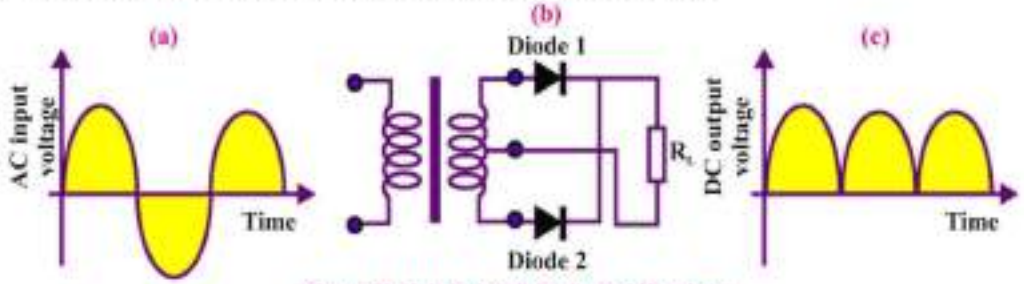


Figure 22.13 (a,b,c) Full-Wave Rectifications

Note that the current through the load is in the same direction for both half cycles of input AC supply. Hence DC output is obtained across the load.

Applications of a Full-wave Rectification:

- Mobile phones, laptops, charger circuits.
- Uninterruptible Power Supply (UPS) circuits to convert AC to DC.
- Our home inverters convert AC to DC.
- LCD, LED TVs.

22.2.6 Function and uses of light emitting diode (LED), Photodiode, Photo voltaic cell:

1. Light Emitting Diode (LED):

A light-emitting diode (LED) is a special type of junction in which current flows, when it is activated in a forward direction as shown in figure 22.14.

In this process, electrons move from the negative side to the positive side, combining with holes emits a photon with energy similar to the material's band gap.

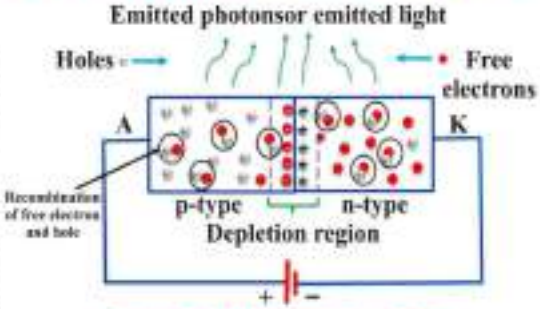


Figure 22.14 Light emitting diode

Silicon diodes don't work well for this, but compounds like gallium and arsenic create efficient LEDs. Gallium arsenide (GaAs), with a crystal structure similar to silicon, is often used..

LED bulbs replace traditional lighting in various applications such as flashlights, streetlights, traffic signals, car brake lights, billboards, and LCD screens. LEDs, also known as solid-state lighting, last longer (50000 hours vs 2000 hours for regular bulbs), are more efficient, and durable.

DO YOU KNOW?

Doping GaAs with group VI atoms like Se and group II atoms as acceptors enhances its properties. GaAs has a 1.42 eV energy gap, producing near-infrared photons (870 nm wavelength), making it suitable for infrared LEDs in remote controls for devices like TVs, DVD players, and car door locks.

2. Photodiode:

A photodiode is a type of light detector that converts light into current or voltage. It includes optical filters, built-in lenses, and surface areas.

Photodiodes are often used in reverse bias, where a voltage encourages the flow of photocurrent, maximizing sensitivity. These detectors are usually made of semiconductor materials like silicon, which can absorb photons of light.

When light particles (photons) hit the semiconductor material, they transfer energy to electrons, creating electron-hole pairs. This process generates an electric current proportional to the incident light's intensity.

In simple terms, brighter light results in a higher current from the photodiode.

Photodiodes find widespread use in various applications, including:

- **Optical Communication:** Photodiodes are employed in optical communication systems, such as fiber optics, to detect and convert transmitted light signals into electrical signals.
- **Light Sensors:** They are used in electronic devices like cameras, light meters, and automatic lighting systems to sense ambient light levels.
- **Barcode Readers:** Photodiodes are often used in barcode scanners to detect the reflected light from the barcode.
- **Medical Devices:** In some medical instruments, photodiodes are used for tasks like measuring oxygen levels in blood.

3. Photo Voltaic Cells:

Solar cell or photovoltaic cell is heavily doped P-N junction diodes used to convert sunlight into electric energy.

If the energy of incident photon is greater than the band gap energy, it creates electron-hole pairs, as shown in figure 22.15. That absorbed photon excites an electron from the valence band and produce a current, when connected to an external circuit, work as source of emf and power. The basic element of a photovoltaic cell is a

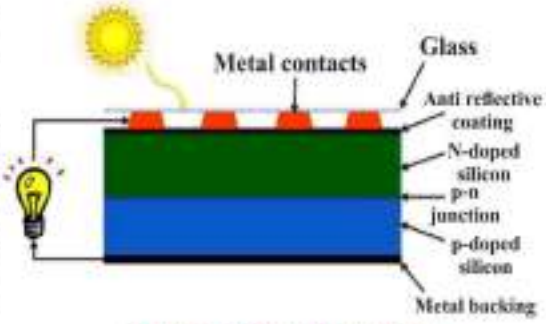


Figure 22.15 Photo Voltaic Cells

semiconductor material, usually made of silicon. Silicon is chosen for its widespread availability. Other semiconductor materials, such as cadmium telluride or copper indium gallium selenide, are also used in different types of solar cells.

A typical silicon *P-N* junction may produce about 0.6 V. Many are connected in series to produce higher voltage. Such series strings are connected in parallel within a photovoltaic panel i.e., solar panel. A good photovoltaic panel can have an output of perhaps 50 W/m^2, averaged over day and night, sunny and cloudy.

Uses of Photodiodes:

- **Camera:** Light meters, Automatic Shutter Control, Photographic Flash Control
- **Medical:** CAT Scanners, X-Ray Detection, Pulse Oximeters, Blood Particle Analysis
- **Automotive:** Headlight Dimmer, Twilight Detectors
- **Communication:** Fiber Optic Link, Optical Remote Control
- **Industry:** Bar Code Scanners, Light Pen, Encoders, Surveying Instruments, Copiers-Density of Toner.

22.3 Transistors and their characteristics:

A transistor is a semiconductor device that amplifies or switches electronic signals by controlling the flow of current between its three layers: base, collector, and emitter. It acts as a gatekeeper, allowing or blocking the flow of current between the collector and emitter, based on the voltage applied to the base.

Both *N-P-N* and *P-N-P* transistors can be made, and they are shown schematically in figure 22.16.

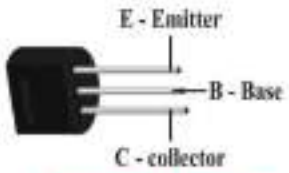


Figure 22.16 (a) Transistor

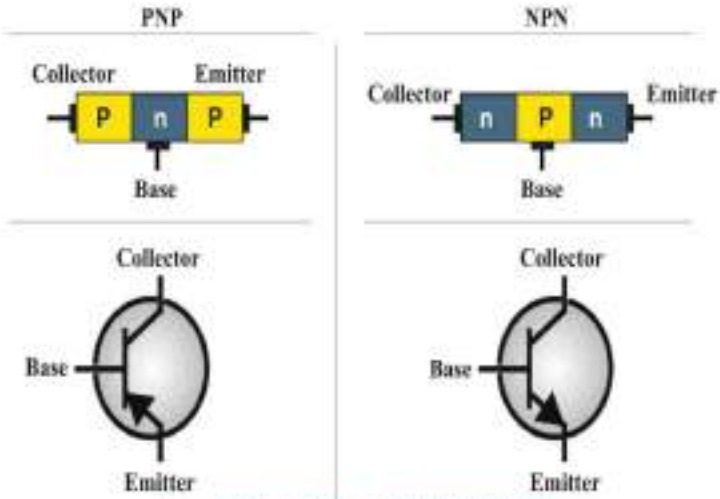


Figure 22.16 (b) PNP and (c) NPN JUNCTION

The arrow is always placed on the emitter; it indicates the direction of current flow in normal operation.

There are three main characteristics listed below:

1. Input Characteristics
2. Output Characteristics
3. Constant-Current Characteristics.

22.3.1 Distinguish between *N-P-N* and *P-N-P*:

N-P-N	*P-N-P*
It contains two N-type semiconductors.	It contains two P-type semiconductors.
Its current is directed outwardly.	Its current is directed inwardly.
In this transistor combination, electrons are majority current carriers.	In this transistor combination, holes are majority current carriers.
Its current direction is emitter to base.	Its current direction is base to emitter.
It has holes as minority current carriers.	It has electrons as minority carriers.
It's switched on when electrons enter the base.	It's switched on when holes enter the base.
As current reduces at base, transistor does not function across the collector terminal and switches off.	When current is available at base, then transistor switches off.

22.3.2 Operation of transistors:

Transistors are categorized as PNP and NPN. For simplicity the NPN transistor is taken as shown in figure 22.17 (a). Its action has two requirements:

1. The base-emitter junction is forward biased so that a current I_B is generated. Once this junction is conducting, $V_{BE} \cong 0.6$ V
2. The base-collector junction is reverse biased.

Transistor action then translates into relationship

$$I_C = \beta I_B \quad \text{.....22.1}$$

Where β is the current gain, which is typically 100. For completeness, the emitter current is then given by

$$I_E = I_B + I_C = (1+\beta)I_B \quad \text{.....22.2}$$

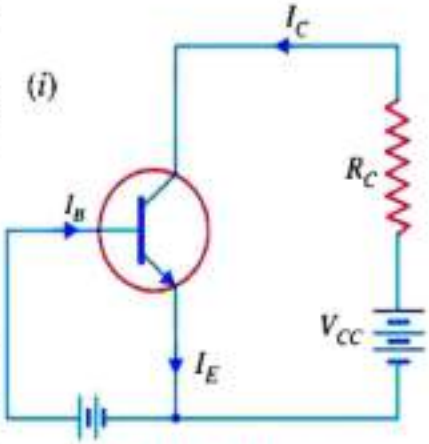


Figure 22.17 (a) NPN transistor

Figure 22.17 graphs a set of characteristics curves of the transistor. Those graphs show the collector current Ic as function of the collector-emitter voltage V_{CE}, for different values of the base current I_B. When the transistor-action conditions are not met, the transistor is off and Ic =0. This situation is called ***cutoff.***

the load resistance (R_L) is connected in series with the collector. The current gain is represented by Beta (β),which is the ratio of I_C to I_B. For the common emitter setup,

$$I_E = I_C + I_B \ldots\ldots\ldots\ldots 22.3$$

where the ratio of I_C to I_E is known as Alpha (α).It's important to note that Alpha is always less than one.

Due to the physical construction of the transistor, any small change in the base current (I_B) leads to a much larger change in the collector current (I_C). This means that slight variations in the current flowing through the base can effectively control the current in the emitter-collector circuit.

Typically, Beta values range between 20 and 200 for most transistors. For instance, if a transistor has a Beta value of 100, one electron flows from the base terminal for every 100 electrons flowing between the emitter-collector terminals.

The mathematical relationship between Alpha (α) and Beta (β) expresses the current gain of the transistor. $\alpha = \frac{I_C}{I_E}$ *and* $\beta = \frac{I_C}{I_B}$ $\therefore I_C = \alpha I_E = \beta I_B$

$$As \quad \alpha = \frac{\beta}{\beta + 1} \quad and \quad \beta = \frac{\alpha}{1 - \alpha}$$

$$I_E = I_C + I_B \ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots\ldots 22.4$$

Where: "I_C" is the current flowing into the collector terminal, "I_B" is the current flowing into the base terminal and "I_E" is the current flowing out of the emitter terminal.

This type of bipolar transistor configuration has greater input impedance, current and power gain than that of the common base configuration but its voltage gain is much lower. The common emitter configuration is an inverting amplifier circuit. This means that the resulting output signal has a 180° phase-shift with regards to the input voltage signal.

Worked Example 22.1

Find Alpha (α) rating of the transistor shown in diagram. Hence determine the value of I_C using both α and β=49rating of the transistor.

Solution:

Step 1: Write down the known quantities and quantities to be found.

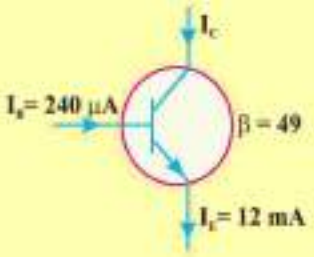
I_B=240μA, I_E=12mA and β=49

I_C=?

Step 2: Write down the formula and rearrange if necessary.

$$\alpha = \frac{\beta}{\beta + 1} = \frac{49}{49+1} = \quad \mathbf{0.98}$$

The value of I_C can be found by using either α or β as under



$I_C = \alpha \times I_E$

$I_C = 0.98 \times 12$ mA

I_C= 11.76 mA

$I_C = \beta \times I_B$

$I_C = 49 \times 240\mu A$

I_C= 11.76 mA

22.3.4 Operation Principles and I-V Characteristics and Methods of Transistor Biasing:

Bipolar Junction Transistor (BJT):

Operation Principles:

1. **Emitter, Base, and Collector:**
   - The three layers of a BJT are called the emitter (E), base (B), and collector (C).
   - The flow of current in an N-P-N transistor is from the emitter to the collector, and in a P-N-P transistor, it's from the collector to the emitter.
2. **Transistor Action:**
   The operation of a BJT involves the injection of minority carriers (electrons in N-P-N or holes in P-N-P) from the emitter into the base region. This controls the majority carrier flow from the collector to the emitter.

I-V Characteristics:

Transistors use current-voltage characteristics to show how input voltage relates to output current. This helps understand how transistors amplify or switch signals.

Active Region:

Here, the transistor amplifies signals. A small current at the base-emitter junction (I_B) causes a larger current from collector to emitter (I_C). The graph shows a straight line, illustrating signal amplification.

Saturation and Cutoff:

Beyond the active region, these regions appear. In saturation, the transistor is ON, allowing maximum collector current. In cutoff, it's OFF, with minimal collector current.

Proper biasing keeps transistors in the active region for linear amplification. Understanding these characteristics is crucial for designing circuits for signal processing, amplification, and switching. Understanding these I-V characteristics is essential for designing circuits that utilize transistors for signal processing, amplification, and switching applications.

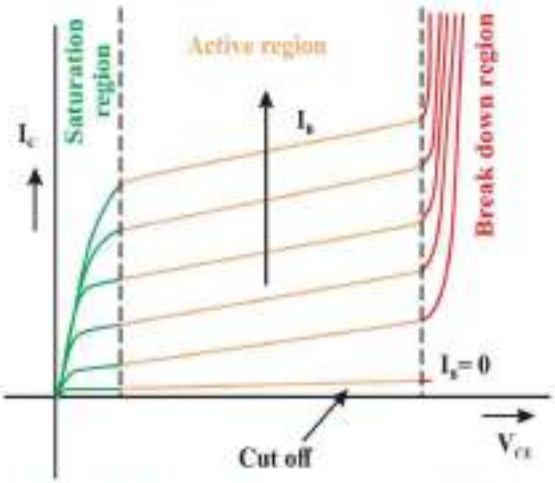


Figure 22.20. Active Region: Saturation and Cutoff

Biasing Methods:

Biasing is the process of applying external voltages to a transistor to establish a desired operating point for proper amplification or switching. There are several biasing methods for transistors. Here are some common biasing methods:

Bipolar Junction Transistor (BJT) Biasing Methods:

1. **Fixed Bias (Base Bias):**
   - Connects a resistor between the base and the power supply to establish a fixed voltage at the base.
   - Simple but not very stable.
2. **Emitter Bias (Emitter Stabilized Bias):**
   - Connects a resistor from the emitter to the power supply, stabilizing the operating point.
   - More stable than fixed bias.
3. **Collector Feedback Bias:**
   - Combines features of fixed bias and emitter bias for improved stability.
   - Uses a resistor network to provide feedback.
4. **Voltage Divider Bias:**
   - Utilizes a resistor divider network to bias the base and provide stability.
   - More stable than fixed bias but less stable than emitter bias.

22.3.5 The transistors as a switch:

A transistor can be used as a switch in electronic circuits, where it functions to either allow or block the flow of current. There are two main configurations for using a transistor as a switch: the NPN (negative-positive-negative) and PNP (positive-negative-positive) configurations.

NPN Transistor as a Switch:

The transistor Q_1 in figure 22.22 shows how to control output current with input.

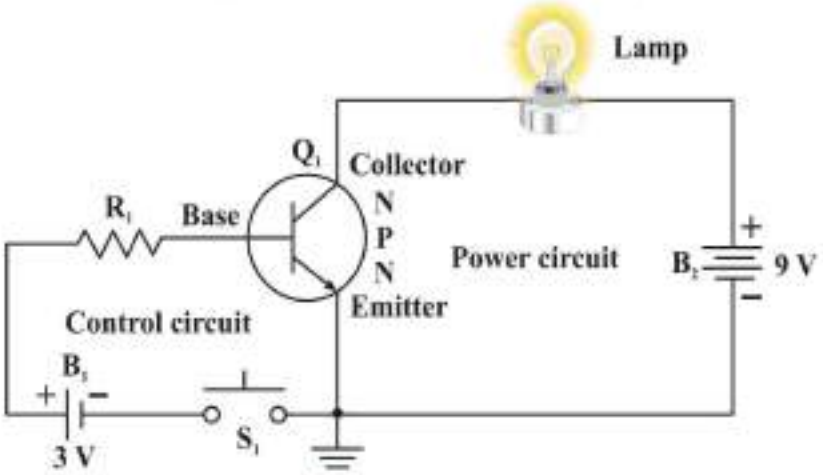


Figure 22.22 transistors as a switch

Here are key points to note:

1. **Off State:**

Normally, Q_1 allows no output current unless we apply forward voltage to its base-emitter circuit.

2. Forward Voltage:

The amount of output current is controlled by the forward voltage that controls base current. In figure 22.22:

- The input control circuit determines base current.
- The output current is collector current for the power circuit.
- Q_1 is an NPN transistor, needing a positive VBE for forward voltage.
- The emitter is common to both input control and power circuits.
- Common-emitter (CE) circuit is the most common transistor arrangement.

The base-emitter junction of Q_1 in figure 22.22 can be forward biased by battery B_1. Switch S_1 must be closed to apply forward voltage. Battery B_2 provides reverse voltage to the collector of Q_1. Reverse polarity means the collector is more positive than the base.

When switch S_1 is open:

- No current flows in the base-emitter or control circuit because no forward voltage is applied.
- Resistance from emitter to collector of the transistor is very high.
- No current flows in the power circuit, and the lamp doesn't light up.

When switch S_1 is closed:

- A small current flows in the control circuit.
- R_1 limits current in the base circuit.
- Resistance from emitter to collector of Q_1 decreases.
- A large current flows in the power circuit, lighting up the lamp.

Opening switch S_1 in the control circuit turns off the lamp in the power circuit because resistance from emitter to collector of Q_1 increases again, almost to infinity.

Transistor as an Amplifier:

Consider the circuit shown in figure 22.23. The dc operating voltages are shown in figure 22.23(a), without any input signal. The dc biasing levels are set by the feedback resistor R_2. The NPN transistor is biased so that the collector-to-emitter voltage Vc is half of the supply voltage. For the supply voltage of 10 v, therefore, the collector voltage is set at one-half the total, or +5 V. The 0.7 V at the base is partially turning on the transistor. The transistor acts as an amplifier when in this partially turned-on condition. It is the amount of dc forward bias that determines the operating level of the transistor.

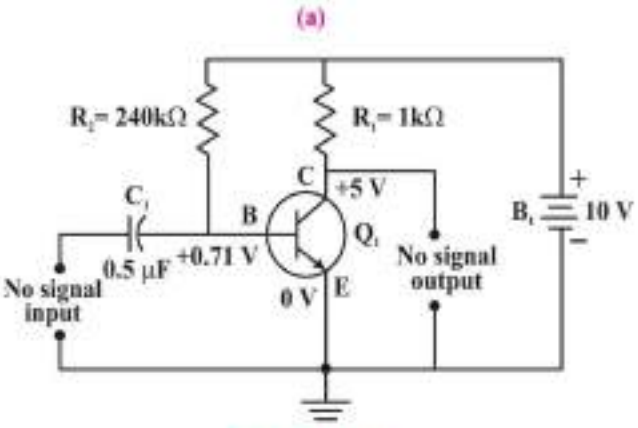


Figure 22.23 (a)
Transistors as an Amplifier

An input signal has been added to the amplifier in figure 22.23 (b), Input is coupled to the base by C_1. Amplified output is taken from the collector. The input signal is 0.02 V_{pp} as measured on an oscilloscope. The measured output signal voltage is 3 V_{pp}.The ac gain of the amplifier, therefore is calculated as

Av = Vout/ V= 3 /0.02 = 150

The output signal (3 V_{pp}) is 150 times greater than the input signal of 0.02 V_{pp}. This amplifier stage is said to have a voltage gain of 150.

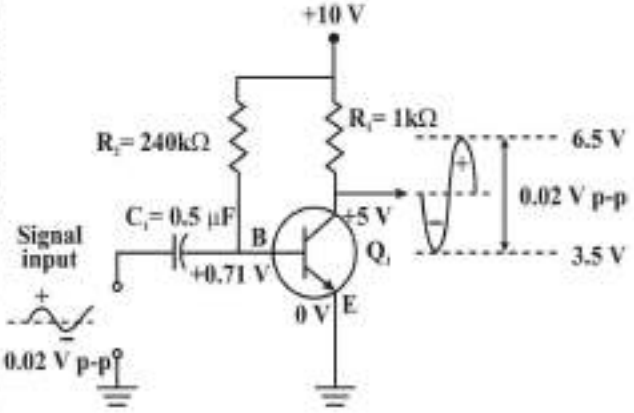


Figure 22.23 (a,b)
Transistors as an Amplifier

Worked Example 22.2

In a common base connection, I_C = 0.95 mA and I_B = 0.05 mA. Find the value of α.

Solution:

Step 1: Write down the known quantities and quantities to be found.

I=0.95mA I_B = 0.05mA R = 470 Ohm α.= ?

Step 2: Write down the formula and rearrange if necessary.

We know $I_E = I_B + I_C = 0.05 + 0.95 = 1mA$

Current amplification factor = $\propto = \frac{I_C}{I_E} = \frac{0.95}{1} = 0.95$

Result: 0.95

22.3.6 Common Collector (CC) Circuit:

In CC configuration, the input signal is applied to the base, and the output signal is taken from the emitter, with the collector terminal serving as the common connection.

Figure 22.24 depicts this setup. Here, when the base-emitter junction is forward biased, a small base current I_B causes a much larger collector current I_C to flow. The emitter current I_E is approximately equal to I_C

since $I_E = I_B + I_C$ and I_B is significantly smaller than I_C

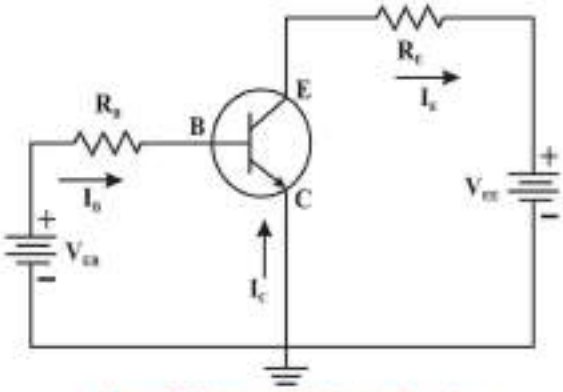


Figure 22.24 common collector circuits

Advantages:

CC amplifiers offer high current gain and low input impedance, making them suitable for applications where a high-resistance input needs to drive a low-resistance output load.

Applications:

Commonly used in impedance matching and signal buffering circuits due to their favorable characteristics.

Common Base (CB) Circuit:

In CB configuration, the input and output signals share the base terminal of the transistor.

Figure 22.25 demonstrates this configuration. Here, the input signal is applied to the emitter, and the output signal is taken from the collector. The base-emitter junction is forward biased to allow transistor operation.

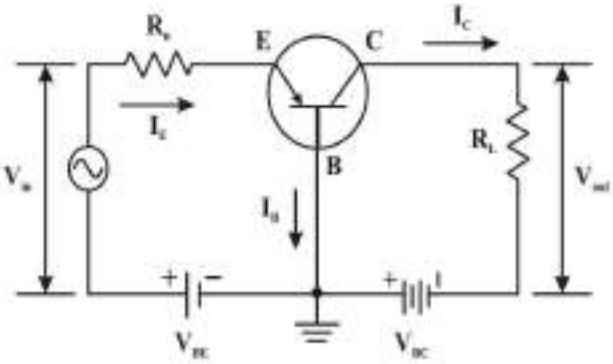


Figure 22.5 Common Base (CB) Circuit

Characteristics:

CB amplifiers provide low input impedance and approximately unity voltage gain. The current gain in CB configuration is expressed as $\frac{I_{out}}{I_{in}}$ or the formula $\frac{I_C}{I_E}$

However, as the base current is extremely small compared to the collector current, the emitter current is therefore approximately equal to the collector current. Thus $I_E \approx I_C$

Applications:

Less commonly used compared to CC and Common Emitter (CE) configurations, CB amplifiers are employed in specific applications where low input impedance and unity voltage gain are advantageous.

22.4 Operational Amplifier:

An operational-amplifier is a **direct coupled high gain amplifier**. It can be operated on both with AC and DC signals. Operational Amplifier, also called as an Op-Amp, is an integrated circuit, which can be used to perform various linear, non-linear, and mathematical operations as shown in figure 22.26.

Op-amps are ideal for DC amplification and are used often in signal conditioning, filtering or other mathematical operations (add, subtract, integration and differentiation).

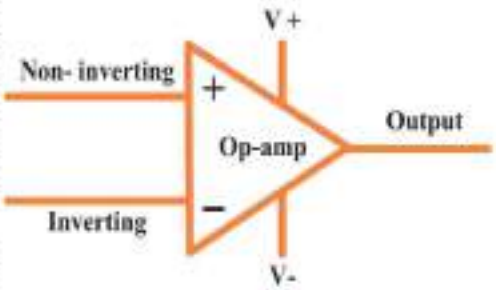


Figure 22.26 Operational Amplifier

22.4.1 Properties of an Ideal Operational Amplifier (Op-Amp):

The ideal op-amp has the following properties:

Characteristics	Ideal Op-Amp
Infinite open-loop voltage gain	An ideal op amp is a device often used as an amplifier. When you input voltage into it, the op amp outputs a amplified voltage. In an ideal scenario, the op amp would provide extremely high gain, essentially infinite, amplifying the signal countless times for maximum gain as needed.
Infinite input resistance (Impedance)	An ideal op amp has super high input impedance, which means it won't load the circuit. If input impedance is low, the op amp draws more current; if it's high, less current is drawn. We aim for high input impedance to avoid disturbing the original circuit by minimizing current pulled from it, ideally with infinite input impedance.
Zero output resistance (Impedance)	In an ideal op amp, the output impedance is zero. This means when the op amp generates a signal, we want it to have zero resistance, ensuring that the maximum voltage goes to the output load. In a circuit, voltage gets divided based on the impedance. Higher impedance means more voltage drop. For the voltage to drop across the output load, the load's impedance should be greater than the op amp's output. That's why, ideally, we aim for zero output impedance in the op amp.
Gain independent of frequency	In an ideal op amp, the produced gain remains constant regardless of the input signal's frequency. This means the amplification stays reliable and consistent across all frequencies.
Infinite bandwidth	The bandwidth of an op-amp is the range of frequencies it amplifies equally. An ideal op-amp amplifies all frequencies, so it should ideally have an infinite bandwidth.
Infinite slew rate	An ideal op-amp should change the output instantaneously as the input is changed. The slew rate of the op-amp is the factor that affects this time delay. An infinite slew rate means there is no time delay.
Zero noise contribution	Any signal includes a small amount of noise. The ideal op-amp does not produce any noise itself, although it will amplify any noise that is present in its input.

22.4.2 Op-Amp as comparator:

The op-amp shown in figure 22.27 is connected to two power supplies. One battery of 9V is connected between the zero-volt line and the +9V positive supply terminal of the op-amp and the other between the zero-volt line and the negative power supply terminal. These batteries are not shown.

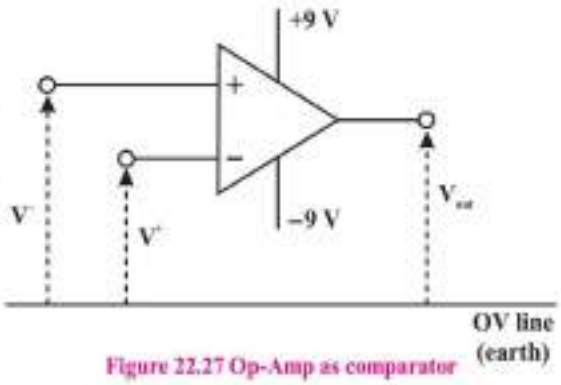


Figure 22.27 Op-Amp as comparator

The output voltage is given by $V_{out} = G_o \times (V^+ - V^-)$, where G_o is the open-loop voltage gain. Notice that all of these voltages are measured with reference to the zero-volt line, which is often connected to earth.

Suppose that $G_o = 10^5$ and that $V^+ = 0.15$ V and $V^- = 0.10$ V. Then equation becomes $V_{out} = 10^5 \times (0.15 - 0.10) = 5000$ V

Certainly, it's impossible. The op-amp is therefore saturated and V_{out} will be close to one of the power supply voltages, in this case 9 V because V^+ is bigger than V^-.

In fact:

- If V^+ is slightly greater in magnitude than V^-, then V_{out} will have magnitude equal to the positive power supply voltage.
- If V^+ is slightly smaller in magnitude than V^-, then V_{out} will have magnitude equal to the negative power supply voltage.

The op-amp is serving as a comparator, comparing V^+ and V^-. If these voltages are slightly different, then the output voltages tell us which one is larger.

22.4.3 Effects of negative feedback on the gain of an operational amplifier:

Negative Feedback:

Negative feedback occurs when a portion of the output signal, like voltage or current, is feedback into the input in a way that opposes or subtracts from the input signal.

Advantages:

1. Reduced Gain:
   Negative feedback decreases the overall gain of the system.
2. Decreased Distortion and Noise:
   It minimizes distortion and noise, leading to a cleaner output signal.
3. Improved Stability:
   The gain becomes more stable and less sensitive to external changes, such as temperature.
4. Increased Bandwidth:
   Enhances the system's bandwidth.
5. Better Impedance:
   Lowers the output impedance and raises the input impedance.

Operation in Electronic Systems:

Feedback is unidirectional, flowing from output to input, ensuring the loop gain (G) remains independent of load and source impedances. A summing point at the input subtracts the feedback signal from the input signal, creating an error signal that drives the system.

Example Circuit (Figure 2.28):

Setup:

Input $V_{in}$is connected to the non-inverting input (+) of an op-amp. The output $V_{out}$is fed back to the inverting input (-).

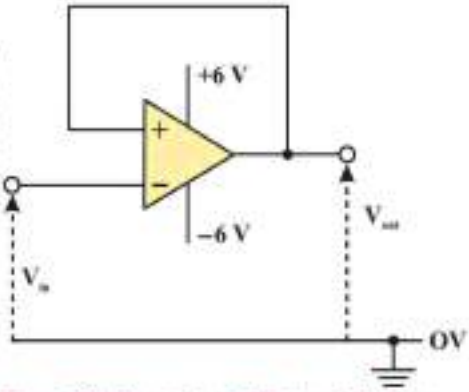


Figure 2.28 Operations in Electronic Systems

Operation:

With an infinite open-loop gain and no saturation, V^- equals V^+.As V_{in} changes, V_{out} adjusts to maintain , $V^- = V^+$.Equation: $V_{out} = G_o \times (V^+ - V^-)$

Simplified: Since$V_{out} = V^-$ and $V_{in} = V^+$ then we have

$V_{out} = G_o (V_{in} - V_{out})$

$V_{out} (1+G_o) = G_o V_{in}$

The closed-loop gain G is given by: $G = \frac{V_{out}}{V_{in}} = \frac{G_o}{1+G_o}$ ……22.5

As G_o is very high about 10^5, there is little difference between G_o and $(1+G_o)$, so the closed-loop gain is very nearly 1. This analysis is true as long as the output voltage is smaller than the supply voltage, in this case as long as V_{out} is between -6 v and +6 v.

Practical Example:

A Piezo-electric microphone with high internal resistance can be connected to an op-amp to increase current output while maintaining the same voltage.

Negative Feedback Benefits Summary:

- Less distortion.
- Increased bandwidth.
- The gain is more stable and not affected by changes in temperature, etc.

The output resistance (impedance) can be low and the input resistance high.

22.4.4 The circuit diagrams for both the inverting and the non-inverting amplifier for single signal input:

Inverting operational amplifier:

In inverting operational amplifiers, the op amp forces the negative terminal to equal the positive terminal, which is commonly ground. Therefore, the input current is determined by the V_{IN} / R_1 ratio as shown in Figure 22.29: Inverting Operational Amplifier.

In this configuration, the same current flows through R_2 to the output. Ideally, current does not flow

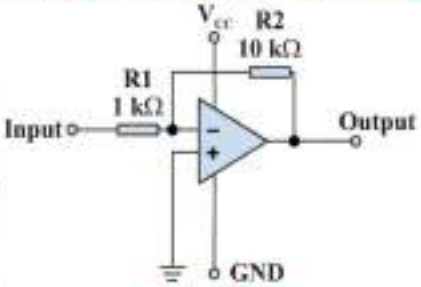


Figure 22.29
circuit diagrams for inverting Amp

into the operational amplifier's negative terminal due to its high Z_{in}. The current flowing from the negative terminal through R_1 creates an inverted voltage polarity with respect to V_{in}. This is why these op amps are labeled with an inverting configuration. Note that the op amp's output can only swing between its positive and negative supplies, so creating a negative output voltage requires an op amp with a negative supply rail. V_{OUT} can be calculated with Equation

$$V_{out} = -\frac{R_2}{R_1} \times V_{in} \ldots\ldots\ldots\ldots\ldots\ldots 22.6$$

Non-inverting operational amplifier:

In a non-inverting amplifier circuit, the input signal from the source is connected to the non-inverting (+) terminal as shown in figure 22.30. Non-Inverting Operational Amplifier

The operational amplifier forces the inverting (-) terminal voltage to equal the input voltage, which creates a current flow through the feedback resistors. The output voltage is always in phase with the input voltage, which is why this topology is known as non-inverting. Note that with a non-inverting amplifier, the voltage gain is always greater than 1, which is not always the case with the inverting configurations. V_{out} can be calculated with Equation:

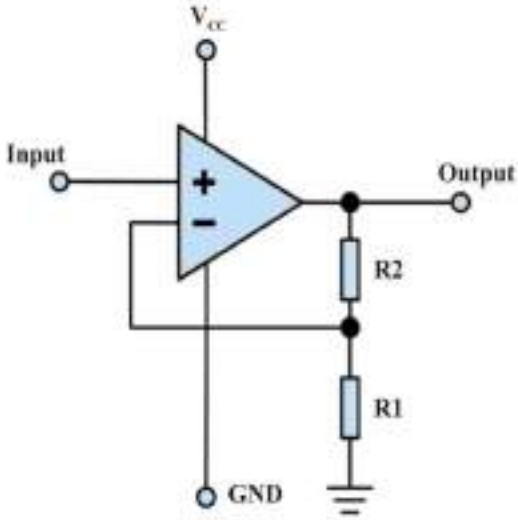


Figure 22.30 circuits for non-inverting amplifier

$$Vout = \frac{1 + R_2}{R_1} \times Vin \ldots\ldots\ldots\ldots\ldots\ldots 22.7$$

22.4.5 Virtual Earth Approximation:

To understand how the inverting amplifier works, you need to understand the concept of virtual earth approximation. In this approximation the potential at the inverting input (-) is very close to 0 V. There are two steps in the arguments:

The op-amps multiplies the difference in potential between the inverting and non-inverting inputs, V^- and V^+, to produce the output voltage V_{out}. Because the open-loop voltage gain is very high, the difference between V^- and V^+ must be almost zero.

The non-inverting input (+) is connected to the zero-volt line so $V^+ = 0$. Thus V^- must be close to zero and the inverting input (-) is almost at earth potential.

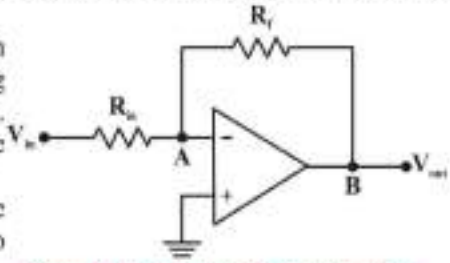


Figure 22.31 virtual earth approximation

The point A is known as a virtual earth as shown in figure 22.31. It cannot actually be 0V but it is very close to 0V. This approximation is true as long as the op-amp is not saturated, and for frequencies where the open-loop voltage gains is high.

Applications:

This approximation is often used in summing amplifier circuits and other applications where it simplifies analysis.

Summing Amplifier Example:

- In a summing amplifier, multiple input voltages can be summed using resistors connected to the inverting input.
- The virtual ground assumption makes it easier to analyze the currents and voltages in the circuit.

Gain of an Inverting Amplifier:

If the current in the input resistor R_{in} is I_{in} and the current in the feedback resistor R_f is I_f, then point P is at 0V:

$$I_{in} = \frac{V_{in}}{R_{in}} \quad and \quad I_f = \frac{V_{out}}{R_f}$$

The input resistance of the op-amp is very high and so virtually no current enters or leaves the inverting input (-) of the op-amp. This means that I_{in} and I_f must be equal in size.
If V_{in} is apositive potential, then the current in two resistors flows from left to right. V_{out} will be negative because the current flows from P, which is at 0 V, to the output connection, which must have a lower voltage than 0 V. Thus

$$I_f = -I_{in} \text{ and } \frac{V_{out}}{Rf} = -\frac{V_{in}}{R_{in}}$$

The gain of the inverting amplifier is thus given by

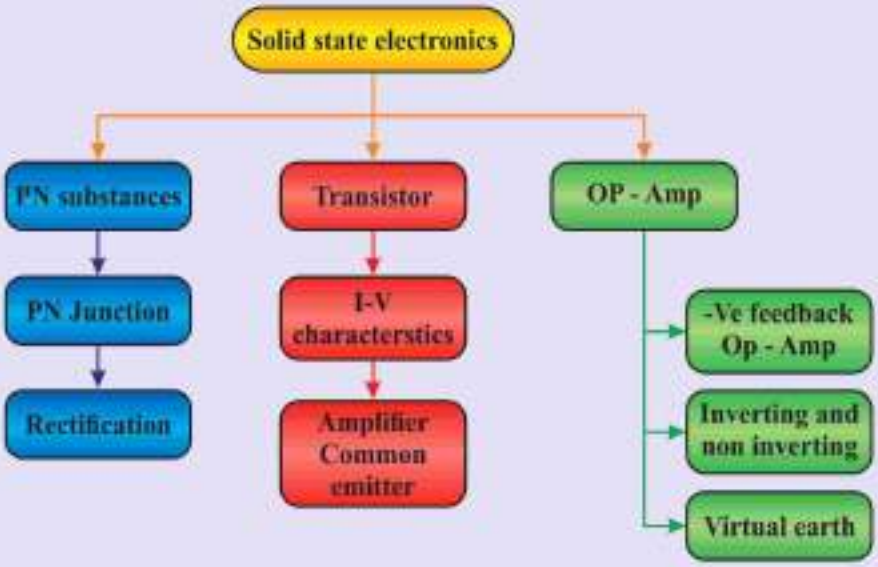
$$G = \frac{V_{out}}{V_{in}} = -\frac{R_f}{R_{in}} \ldots\ldots\ldots\ldots 22.8$$

The negative sign shows that when the input voltage is positive the output voltage is negative and when the input is positive the output is negative. If the input voltage is alternating then there will be a phase difference of 180° or π rad between the input and the output voltages.

SUMMARY

- A *p*-type substance is a semiconductor material which has an excess of positively charged particles called "holes" as its majority charge carriers.
- An *n*-type substance is a semiconductor material that has an excess of negatively charged electrons as its majority charge carriers.
- An intrinsic semiconductor is a pure semiconductor material with no intentional impurities added.
- Doping is the process of intentionally introducing impurities into a semiconductor crystal to modify its electrical properties.
- Doping with Donor Impurities: A small amount of a group V element is introduced into the crystal lattice of the semiconductor. Common donor impurities include phosphorus (P) or arsenic (As).
- Doping with Acceptor Impurities: A small amount of a group III element is introduced into the crystal lattice of the semiconductor. Common acceptor impurities include boron (B) or gallium (Ga).
- P-N Junction is formed when the *p*-type and *n*-type semiconductors are joined, is called as *p-n* junction.
- P-N Junction Diode is a semiconductor device that is formed through this method and used in allowing the flow of electric current in one direction and blocking in the opposite.
- Forward Bias: The *P-N* junction is forward-biased when the *P*-type is connected to the positive terminal of the battery and the *N*-type is connected to the negative terminal.
- Reverse Bias: The *P-N* junction is reverse biased when the*P*-type is connected to the negative terminal of the battery and the *N*-type is connected to the positive side.
- Rectification is the process of converting an alternating current (AC) waveform into a direct current (DC) waveform, i.e., creating a new waveform that has only a single polarity.
- A light -emitting diode (LED) is a *P-N* junction in which current flows as it becomes forward biased.
- A photodiode is one type of light detector which uses to convert the light into current or voltage based on the mode of operation of the device.
- Photo Voltaic Cells or Solar cells, also called photovoltaic cells, are rather heavily doped *p-n* junction diodes used to convert sunlight into electric energy.
- Transistor: The Bipolar Junction Transistor was invented in 1948 by J. Bardeen, W. Shockley and W. Brattain.
- It consists of a crystal of one type of doped semiconductor sandwiched between two of the opposite type.

CONCEPT MAP



EXERCISE

Section (A): Multiple Choice Questions (MCQs)
Choose the correct answer:

1. A semiconductor is an element with a valence electron _____.
   (a) Four (b) Eight (c) Two (d) One
2. A pure semiconductor is known as
   (a) Extrinsic (b) Intrinsic (c) Transistor (d) Diode
3. An acceptor atom is also called
   (a) Penta-valent atom (b) Trivalent atom
   (c) Minority carrier (d) Majority carrier
4. With which one of the following elements silicon should be doped so as to give p-type of semiconductor?
   (a) Germanium (b) Arsenic (c) Selenium (d) Boron
5. For a full-wave rectifier, the output frequency
   (a) Equals one-half the input frequency (b) Equals the line frequency
   (c) Equals two times the input (d) Is three times the line frequency

6. LED construction needs a semiconductor material is:
   (a) Silicon (b) Germanium (c) Gallium (d) Gallium arsenide
7. The frequency of a half-wave signal is
   (a) Twice the line frequency (b) Equal to the line frequency
   (c) One-half of the line frequency (d) One-fourth the line frequency
8. The voltage gain of an emitter follower circuit is
   (a) High (b) Low (c) Very high (d) Moderate
9. What is also called as the conventional amplifier?
   (a) Common-collector circuit (b) Emitter follower circuit
   (c) Common base circuit (d) Common emitter circuit
10. An op-amp with negative feedback provides _____ output parameter.
   (a) Gain (b) Bandwidth
   (c) Input-output impedance (d) Amplified

Section (B): CRQs (Short Answered Questions):

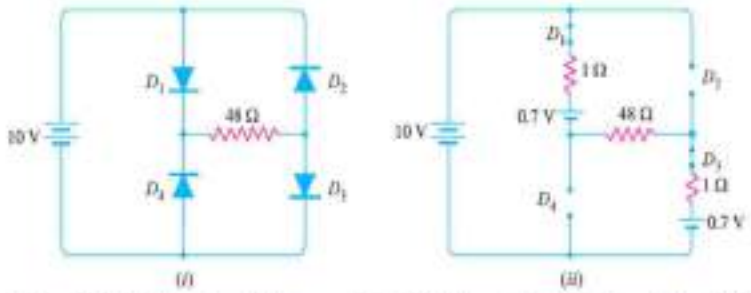
1. Describe a *p-n* junction (diode) and how holes and electrons are produced in semiconductor?
2. Define and distinguish between p-n-p & n-p-n transistors?
3. Explain common-base and common collector configurations?
4. Describe the operations of transistors.
5. Explain the use of transistors as a switch and an amplifier (common-emitter).
6. How would you understand the effects of negative feedback on the gain of an operational amplifier?
7. Draw and briefly describe the circuit diagrams for both the inverting and the non-inverting amplifier for single signal input?
8. Derive an expression for the gain of inverting amplifiers by using virtual earth approximation.
9. Describe the properties of an ideal operational amplifier.

Section (C): ERQs (Long Answered Questions):

1. Define Intrinsic (pure) and doped semiconductors. And How the N-type and P-type semiconductors are produced
2. Describe a P-N junction (diode) discus its forward and reverse biasing.
3. Describe the I-V characteristic curves of P-N junction.
4. Define rectification and describe the use of diodes for half and full wave rectifications.
5. Describe the function and use of LED, Photodiode and Photo voltaic cell Describe the function and use of LED, Photodiode and Photo voltaic cell.
6. Explain the use of transistors as a switch and an amplifier (common-emitter).
7. Describe the properties of an ideal operational amplifier.

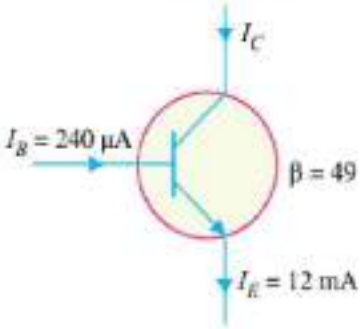
Section (D): **Numerical:**

1. A Ge diode has a voltage drop of 04 V when 12 mA flow through it. If the same 470 Ohm is used, what battery voltage is needed? **(6.04 V)**
2. A semiconductor diode laser has a peak emission wavelength of 1.55 μm. Find its band gap in eV. **(0.8 eV)**
3. Calculate the current through 48 Ω resistor in the circuit shown in Fig. (i). Assume the diodes to be of silicon and forward resistance of each diode is 1 Ω. **(8.6 V, 50 Ohm, 172 mA)**



4. Find the voltage V_A in the circuit shown in Fig.(i). Use simplified model. (19.7 V)

5. In a common base connection, $I_E = 1mA$, $I_C = 0.95mA$. Calculate the value of I_B. **(0.05 mA)**
6. Find the value of β if (i) α = 0.9 (ii) α = 0.98 (iii) α = 0.99. **(9, 49, 99)**
7. Calculate I_E in a transistor for which β = 50 and I_B = 20 μA. **(1.02 mA)**
8. Find the α rating of the transistor shown in Fig. Hence determine the value of I_C using both α and β rating of the transistor. **(0.98, 11.76)**



Introduction:

Digital electronics is a branch of electronics that deals with the manipulation of digital signals, represented as binary digits (0s and 1s).

Unlike analog electronics, which uses continuous signals, digital electronics processes discrete signals, enabling precise control and storage of information. This field encompasses the design, analysis, and application of digital circuits and systems, forming the foundation for modern computing, communication, and information processing technologies.

23.1 Digital signal levels:

In digital electronics, signal levels are represented in various ways. Here are common digital signal levels:

1. Low Level (0):

- This represents the binary digit 0.
- In terms of voltage, it is associated with a lower voltage level.
- Often referred to as "low" or "logic 0."

2. High Level (1):

- This represents the binary digit 1.
- In terms of voltage, it is associated with a higher voltage level.
- Often referred to as "high" or "logic 1."

3. Threshold Level:

- The threshold level is the voltage level that separates low and high states.
- Signals below this threshold are interpreted as 0, and signals above it are interpreted as 1.
- The threshold level helps define the noise margin and ensure reliable signal interpretation.

4. Logic Levels:

- Depending on the technology and the system, different logic levels might be used.

5. Swing or Voltage Range:

- The difference in voltage levels between low and high states is often referred to as the swing or voltage range. For example, voltage varies from -5V to +5V will have a swing or voltage range of 10V.
- A larger voltage swing can enhance noise immunity and signal reliability.

DO YOU KNOW?

Noise immunity is a measure of how well a logic circuit can withstand unwanted electrical signals or noise without reacting to them.

DO YOU KNOW?

If we deal with mathematics, then the binary system, using 1 or 0, will be employed. If we are dealing with voltage in a circuit, then HIGH or LOW can be utilized.

23.1.2 Understanding Digital Signal Levels:

In digital electronics, the 'high' and 'low' states are often symbolically represented as '1' and '0,' respectively. This binary notation helps describe the states of open and closed circuits.

- **High State ('1'):** When a digital signal is in a 'high' state, it is typically associated with a closed circuit or a voltage level close to the maximum specified value. The symbolic representation for this state is '1'. It signifies the 'on' or 'active' state in many digital systems.
- **Low State ('0'):** Conversely, the 'low' state is linked to an open circuit or a voltage level close to the minimum specified value. It is symbolically represented as '0'. This state signifies the 'off' or 'inactive' state in digital systems.

High and Low represent the **signal levels** in the circuit. Digital waveforms consist of voltage levels that are changing back and forth between the HIGH and LOW states. These can also be described as the one that is composed of series of pulses, or a pulse train as shown in Figure 23.1.

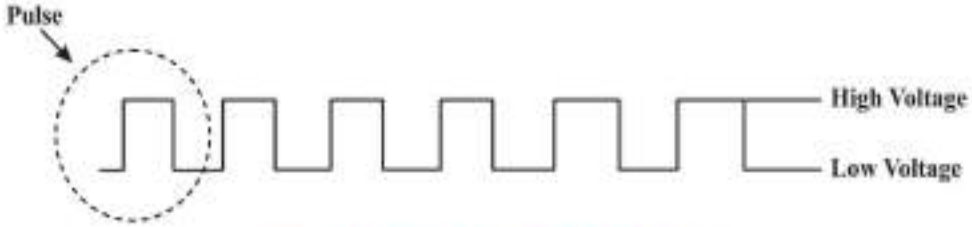


Figure 23. 1 Pulse showing High and Low states

These symbolic representations simplify the expression and interpretation of digital information, allowing for clear communication and logical operations within electronic circuits. The '1' and '0' states form the basic building blocks for encoding and processing binary data in digital systems. This binary representation forms the fundamental language of digital systems, allowing them to process and manipulate data efficiently.

DO YOU KNOW?

The mathematics of binary system is different from ordinary mathematics. The special algebra representing binary system is known as Boolean Algebra.

23.2 Logic gates:

Logic gates are tiny electronic switches that enable decision-making in computers and other devices. They combine to perform logical operations, playing a crucial role in information processing.

23.2.1 Types of Logic Gates:

Logic gates are the basic components of the digital circuit with one output and more than one input. AND, OR, and NOT gates are the basic gates, while NAND and NOR are the universal gates. EX-OR and EX-NOR are the special gates.

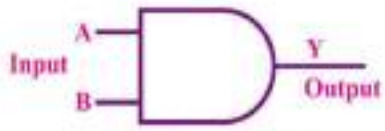


Figure 23. 2 AND gate

Demonstration of Logic Gates in Simple Circuits:

In this part, we are going to have some fun with simple circuits. By utilizing just 2 switches, a lamp and a battery, we explore the behaviors of a 2-input AND gate and a 2-input OR gate. These hands-on examples offer a concrete insight into how logic gates process signals and make logical decisions, forming the basis for more complex digital systems.

DO YOU KNOW?

The XNOR gate produces a high output (1) when an even number of its inputs are high (1).

I. AND Gate Demonstration:

In this demonstration, we will use following components as shown in the circuit diagram:

- 2 switches (Switch A and Switch B)
- 1 lamp
- 1 battery

The presented circuit exhibits two switches (A & B) and a bulb (F) arranged in series with a cell. According to basic circuit theory, the bulb illuminates only when a closed circuit permits the flow of current.

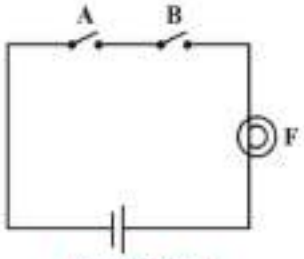


Figure 23. 9(a)
AND gate circuit demonstration

Behavior: The lamp will light up only when both Switch A and Switch B are in the ON position. If either or both switches are OFF, the lamp remains off.

However, in the context of digital electronics, we approach the analysis differently. Instead of focusing on the physical flow of current, we interpret the circuit as making decisions, determining whether the output represents logic 1(the bulb is lit) or logic 0 (bulb not lit). This decision is contingent upon the logical states of the inputs (i.e., the switches), where an open switch corresponds to logic 0 and a closed switch corresponds to logic 1.

A	B	F
0	0	0
0	1	0
1	0	0
1	1	1

Figure 23. 9(b)
AND gate and its truth table

The truth table and the equivalent combination of above circuit in logic gate is shown in figure 23.9.

II. OR Gate Demonstration:

In this demonstration, we will use following components as shown in the circuit diagram:

- 2 switches (Switch A and Switch B)
- 1 lamp
- 1 battery

Figure 23.10 shows a simple circuit which contains two switches in parallel.

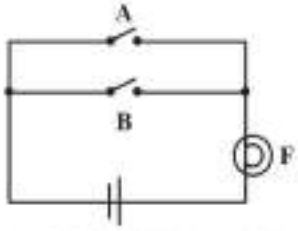


Figure 23. 10(a) OR gate circuit demonstration

Behavior: The lamp will light up if either or both Switch A and Switch B are in the ON position. The lamp remains off only when both switches are OFF.

EXERCISE

Section (A): Multiple Choice Questions (MCQs)
Choose the correct answer:

1. The voltage level, typically associated with a 'low' state in digital electronic is:
   (a) -5 volts (b) 0 volts (c) +5 volts (d) +10 volts
2. In binary representation, the correspond of '0' in digital electronics is:
   (a) Low state (b) High state (c) Open circuit (d) Closed circuit
3. The primary purpose of a truth table in digital electronics is to:
   (a) Determine input voltage
   (b) Measure circuit resistance
   (c) Analyze output states based on input combinations
   (d) Calculate circuit power
4. Which logic gate combination produces an output of '1' when at least one input is '1'?
   (a) AND (b) OR (c) XOR (d) NOT
5. In a circuit with a 2-input OR gate, the status of the lamp if either switch is closed
   (a) OFF (b) ON (c) Blinking (d) Flickering
6. The behavior of 2 input AND gate, when both switches are closed, Lamp is
   (a) OFF (b) ON (c) blinking (d) flickering
7. In a 3-input OR gate, if all inputs are '0'sothe output will be:
   (a) 1 (b) 0 (c) Undefined (d) Both A and B
8. If the input is '1', The output of a NOT gate will be:
   (a) 1 (b) 0 (c) Undefined (d) Both A and B
9. In a 2-input XOR gate, when both inputs are '1' than the output will be
   (a) 1 (b) 0 (c) Undefined (d) Both A and B
10. Which combination of logic gates is equivalent to an XOR gate?
    (a) AND + OR (b) NOT + NOR (c) OR + NAND (d) AND + NAND

Section (B): CRQs (Short Answered Questions):

1. Explain the significance of signal levels in digital electronics and how they are represented in terms of voltage.
2. Describe the binary representation used in digital electronics. How is this representation related to the 'high' and 'low' states?
3. Choose any three logic gate symbols and explain their operations.
4. Create a combination of logic gates that mimics the behavior of an XOR gate.

Section (C): ERQs (Long Answered Questions):

1. Design a simple circuit using a 2-input AND gate, a switch, a lamp, and a battery. Explain how the circuit operates and the conditions under which the lamp will be illuminated.
2. Construct a truth table for a 3-input OR gate. Explain how the truth table represents the relationship between input combinations and the resulting output.

O and O' and a particle P. These observers use frames of reference XYZ and X'Y'Z', respectively. If O and O' are at rest, they will observe the same motion of P. But if O and O' are in relative motion, their observation of the motion of P will be different.

24.1.1 Relative Motion:

Relative motion is a concept that describes the motion of an object in relation to another observer or frame of reference. This viewpoint recognizes that motion is not absolute but depends on the observer's point of view.

Relative Motion: Same direction:

When two objects are moving in the same direction, their relative motion involves the difference in their velocities. The relative velocity is the velocity of one object as seen from the perspective of the other. The formula for relative velocity (v_{rel}) when the objects move in the same direction is given by:

$$v_{rel} = v_1 - v_2$$

where v_1 and v_2 are the velocities of the two objects, respectively.

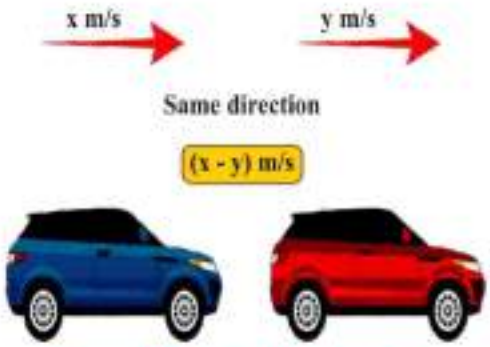


Figure 24.2 moving in same direction

Example:

Consider two cars, A and B, moving on a straight road. If car A has a velocity of 20 m/s, and car B has a velocity of 15 m/s in the same direction, the relative velocity of car B as observed from car A is

v_{rel}=20m/s – 15m/s=5m/s

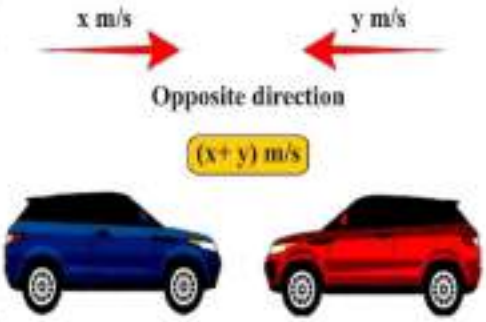


Figure 24.3 moving in different direction

Relative Motion: Opposite direction:

When two objects are moving in opposite directions, their relative motion involves the sum of their velocities. The formula for relative velocity (v_{rel}) when the objects move in opposite directions is given by:

$$v_{rel} = v_1 + v_2$$

Example:

Consider same two cars, A and B. But this time moving in opposite directions. Let's say car A is traveling at 80 m/s, and car B is now approaching it at 50 m/s.For a passenger on car A, car B appears to be moving towards them at relative velocity:

v_{rel}=80 m/s+50 m/s= 130 m/s

24.1.2 Types of Frames of References:

There are two types of frames of reference, Inertial frames of reference an dnon-inertial frames of reference.

Criteria	Inertial frame of Reference	Non-Inertial frame of Reference
Definition	A frame of reference with a constant	A frame of reference with an accelerating motion.
Motion of objects	Objects in uniform motion appear to follow straight lines or constant velocities.	Objects may appear to accelerate or experience fictitious forces.
Newton's first law	Newton's first law (Law of inertia) is valid in this frame.	Newton's first law is not valid due to accelerating motion.
Appearance of forces	Real forces are observed and can be directly measured.	Fictitious forces (e.g., centrifugal force, coriolis force) may appear due to the acceleration of the frame.
Equations of motion	Newton's laws of motion hold true in this frame.	Additional terms or transformations may be required to account of the frame.
Examples	A person inside a moving train, an object in free fall.	A person inside a spinning carousel, an object in circular motion.
Application	Often used in analyzing motion and dynamics in	Important in understanding phenomena in rotating systems, general relativity, and

Self-Assessment Questions:

1. Explain the difference between an inertial and a non-inertial frame of reference.
2. Provide an example of a commonly used frame of reference in everyday life.

24.1.3 Frames of references: Prediction of Motion:

Consider a vehicle moving at a constant speed v = 15 m/s with relative to the ground and a boy standing on the vehicle throws a ball with velocity v = 8 m/s in the same direction as the vehicle's motion. As shown in fig 24.4. If someone on the ground measures the ball's speed, it intuitively seems like it should be 23 m/s (15+8). Conversely, if you throw

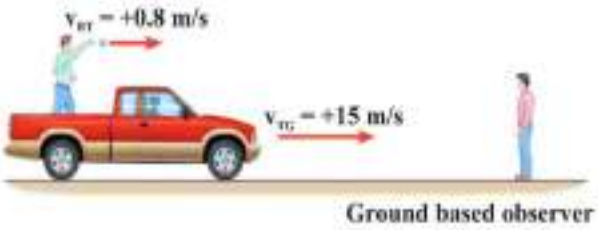


Figure 24.4: Moving and stationary frames of references

the ball in the opposite direction of the vehicle's speed, the ground observer would measure a speed of 7 m/s (15 − 8). Both observers are measuring the speed of the same object, but their relative motion leads to different results.

Now, let's derive a general equation applicable to any speeds. In our example, the vehicle's speed of 15 m/s becomes the relative speed of the two reference frames, denoted as u. The speed of the thrown ball is v' (in the train's frame), and the measured speed on the ground is v. In a more formal way, we can express this relationship as:

$$v = v' + u$$

This equation is essentially the **Galilean transformation** of speeds. But, it is only applicable in mechanics.

Relative Motion: Galilean Transformation Equations:

The Galilean transformation equations describe the relationship between the coordinates of an event in two different inertial frames of references. These equations were formulated by Galileo and are applicable when the relative velocities involved are much smaller than the speed of light. With these equations, we can correctly measure the velocities of a ball in above example. Galileo concluded that:

"The laws of mechanics must be valid in all inertial frames of reference."

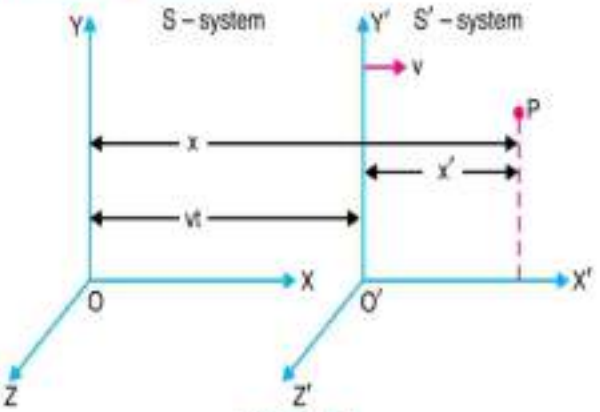


Figure 24. 5
Inertial frames of references: moving and stationary

This is also known as the "Principle of Galilean Relativity".

Let x and x' be two inertial frames (Figure 24.5). Let x be at rest and x' moves with uniform velocity v along the positive X direction. We assume that v << c. Let the origins of the two frames coincide at t = 0. Suppose some event occurs at the point P. The observer O in frame x determines the position of the event by the coordinates x, y, z. The observer O' in frame x' determines the position of the event by the coordinate's x', y', z'. Let the time elapsed at the same rate in both frames, i.e., t = t'. There is no relative motion between x and x' along the axes of Y and Z. Measurements in the X direction made is x frame will be greater than those made in x' frame by the amount v_t, which is the distance x' has moved in the X direction. Therefore,

$$x' = x - vt$$
$$y' = y$$
$$z' = z$$
$$t' = t$$

The set of above equations is known as the Galilean transformation equations.

24.2 Special Theory of Relativity:

Special theory of relativity explains how to interpret motion between different inertial frames of reference. In other words, it deals with the problems in which one frame of reference moves with a constant linear velocity relative to another frame of reference.

24.2.1 The Postulates of Special Relativity:

The special theory of relativity is based on two essential assumptions, commonly known as postulates.

Postulate I (Principle of Relativity):

The laws of Physics have the same form in all inertial frames of reference.

Consider a vehicle moving at a constant speed. Inside the vehicle, a passenger throws a ball straight up into the air. If we ignore any effects from the air, the passenger inside the moving vehicle sees the ball move up and then back down in a straight line, just as it would appear if someone standing still on the Earth threw a ball upwards. This means that the laws of physics that govern the motion of the ball, including gravity and equations for constant acceleration, work the same way whether the vehicle is moving or at rest.

Both observers, the one inside the vehicle and the one on the ground, see the ball go up and come back down. However, they see the path of the ball differently. The person on the ground sees the ball's path as a curved shape (a parabola), while the person in the vehicle sees it as a simple up-and-down motion. As shown in figure 24.6.

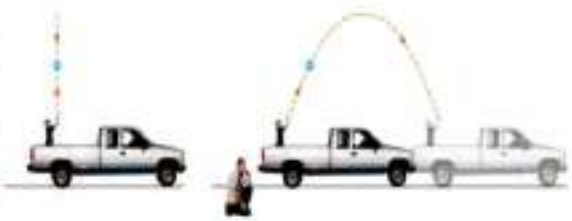


Figure 24.6: Principle of Relativity

Additionally, the person on the ground thinks that the ball has a horizontal component of velocity, which is the same as the vehicle's velocity. For the outside observer, the distance traveled along the parabolic path is longer than the path straight down but the time for the fall is the same. For the outside observer, the average velocity is greater.

Postulate II (Constancy of the speed of light):

The speed of light in vacuum has the same value, $c=3 \times 10^8 m/s$ in all inertial reference frames, regardless of the velocity of the observer or the velocity of the source emitting the light.

Consider again the same example of throwing a ball in moving vehicle. As shown in figure 24.7, But this time, instead of throwing a ball, we shine a flashlight. If we were to do the same thing with light as we did with the ball, common sense might suggest that the

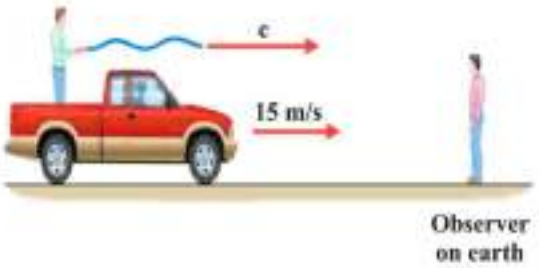


Figure 24.7 Two different frames of reference but speed of light is constant for both frames of references

This indicates the level of stability. It means the clock deviates by one part in 10 trillion (10^{13}) over a certain period. In other words, the clock's timekeeping is extremely precise, with an error margin of just 0.0000000001%.

Worked Example 24.1

The period of a pendulum is measured to be 3.00 s in the inertial frame of the pendulum. What is the period as measured by an observer moving at a speed of 0.950c with respect to the pendulum?

Solution:

Step 1: Write down the formula and rearrange if necessary.

Using the equation of mass variation:

$$t = \frac{t_o}{\sqrt{1 - \frac{v^2}{c^2}}}$$

Step 2: Substitute the proper time and relative speed in the equation of time dilation:

$$t = \frac{t_o}{\sqrt{1 - \frac{v^2}{c^2}}} = \frac{3}{\sqrt{1 - \frac{0.95c^2}{c^2}}} = 9.61s$$

Result: The moving observer considers the pendulum to be moving, and moving clocks are observed to run more slowly: while the pendulum oscillates once in 3 s for an observer in the rest frame of the clock, it takes nearly 10 s to oscillate once according to the moving observer.

Worked Example 24.2

A starship is measured to be 125 m long while it is at rest with respect to an observer. If this starship now flies past the observer at a speed of 0.99c, what length will the observer measure for the starship?

Solution:

Step 1: Write down the formula and rearrange if necessary.

Using the equation of mass variation:

$$L = L_o\sqrt{1 - \frac{v^2}{c^2}}$$

Step 2: Substitute the proper length and relative speed in the equation of length contraction:

$$L = L_o\sqrt{1 - \frac{v^2}{c^2}} = 125 \times \sqrt{1 - \frac{0.99c^2}{c^2}} = 17.6m$$

Result: $L = 17.6m$

Worked Example 24.3

At what speed will an object's relativistic mass be twice its rest mass?

Solution:

Step 1: Write down the formula and rearrange if necessary.

Using the equation of mass variation: $m = \frac{m_o}{\sqrt{1-\frac{v^2}{c^2}}}$

Step 2: Where its relativistic mass $m = 2m_o$. Then velocity of the object will be:

$$2m_o = \frac{m_o}{\sqrt{1-\frac{v^2}{c^2}}} \quad = 2 \times \sqrt{1-\frac{v^2}{c^2}} = 1$$

Step 3: Squaring both sides:

$$1 - \frac{v^2}{c^2} = \frac{1}{4}$$

$$1 - \frac{1}{4} = \frac{v^2}{c^2}$$

$$v = \frac{\sqrt{3}c}{2}$$

Result: This speed is approximately 86.6% of the speed of light.

24.3 General Theory of Relativity:

Einstein worked on the fact that acceleration produced the same effect as gravitation. For example, if you were in an accelerating spaceship, or just an elevator, you could not tell if the force on you was from inertia or gravitation. While Newton's Law of Universal Gravitation works well for ordinary gravitational fields, it is inaccurate when the gravitational intensity is high.

27.3.1 General Theory of Relativity:

Albert Einstein's general theory of relativity is a new way to explain gravity. According to Newton, tossed balls curve because of a force of gravity. According to Einstein, tossed balls and light don't curve because of any force, but because the space time in which they travel is curved. General relativity describes how gravity is caused by the curvature of space-time.

The general theory of relativity is based on the following two postulates.

Spacetime is a four-dimensional continuum that includes three dimensions of space and one dimension of time. Imagine it as a two-dimensional rubber sheet. If you place a heavy object, like a bowling ball, on the sheet, it will cause the sheet to deform or bend around the object. Now, if you roll a smaller ball across the sheet, it will move along the curved paths created by the heavy object. This is analogous to how planets and other celestial bodies move along the curved space-time around massive objects like the Sun. As illustrated in Figure 24. 14.

SUMMARY

- ✓ Special Reactivity applies to objects moving at constant velocity relative to each other. It introduces the concept of reference frames and explains how the laws of physics are the same in all inertial reference frames.
- ✓ Two important postulates of Special Relativity are:
  - ➢ The laws of physics are the same in all inertial frames of reference.
  - ➢ The speed of light in a vacuum is the same for all observers regardless of their relative motion or the motion of the source of the light.
- ✓ Consequences of Special Relativity include:
  - ➢ Relativity of simultaneity: Two events that are simultaneous in one frame of reference may not be simultaneous in another frame moving relative to the first.
  - ➢ Time dilation: Time intervals between events appear longer for objects in motion relative to an observer at rest.
  - ➢ Mass variation refers to the change in the mass of an object as its speed approaches the speed of light. According to relativity, the mass of an object increases with its velocity, becoming infinite as it reaches the speed of light.
  - ➢ Length contraction: Objects in motion appear shorter in the direction of their motion relative to an observer at rest.
  - ➢ Mass-energy equivalence: Mass and energy are interchangeable $E=mc^2$.
- ✓ General Relativity applies to objects under any motion, including acceleration. It explains gravity as a curvature of space-time caused by mass and energy.
- ✓ Two postulates of General Relativity are:
  - ➢ The laws of physics may be expressed in equations having the same form in all frames of reference, regardless of their states of motion.
  - ➢ There is no way for an observer in a closed laboratory to distinguish between the effects produced by a gravitational field and those produced by an acceleration of the laboratory itself.
- ✓ Consequences of General Relativity include:
  - ➢ Space-time: Space and time are unified into a single four-dimensional fabric.
  - ➢ Gravitational red-shift: Light coming from a strong gravitational field has a longer wavelength (red-shifted).
  - ➢ Black holes: Extremely massive objects whose gravity is so strong that not even light can escape.
  - ➢ Gravitational lensing: Light bends as it travels through curved spacetime, which can distort, diminish or magnify distant objects.
  - ➢ Expansion of the universe: General relativity predicts the universe is expanding, which has been confirmed by astronomical observations.

EXERCISE

Section (A): Multiple Choice Questions (MCQs)
Choose the correct answer:

1. The General Theory of Relativity was a new way of understanding of
   (a) The speed of light (b) gravity
   (c) mass (d) force
2. An object at rest has a mass of 1 kg. What is its mass when it is moving at a speed of 0.9c?
   (a) Infinite (b) 1.2 kg (c) 2.3 kg (d) 1 kg
3. The equivalence of principle in general relativity is:
   (a) The equivalence of inertial and gravitational mass
   (b) The equivalence of electric and magnetic fields
   (c) The equivalence of space and time
   (d) The equivalence of matter and energy
4. Which of the following phenomena is NOT predicted by special relativity?
   (a) Time dilation (b) Length contraction
   (c) Gravitational waves (d) Relativistic mass increase
5. What does the equivalence principle state about acceleration and gravity?
   (a) They are completely different forces (b) They are indistinguishable for an observer
   (c) Gravity is stronger than acceleration (d) Acceleration cancels out gravity
6. Light passing near a massive object like a star will bend due to:
   (a) Gravitational lensing (b) refraction
   (c) reflection (d) diffraction
7. What is the main reason astronomers cannot directly observe black holes?
   (a) They are too small (b) They deflect light
   (c) They are too far away (d) their immense gravity traps light
8. Objects cannot exceed the speed of light because:
   (a) Their mass becomes infinite (b) Their length becomes zero
   (c) Time slows down to zero (d) they lose all energy
9. Why is the Galilean transformation not valid at high speeds?
   (a) It cannot handle accelerating frames
   (b) It violates the constancy of the speed of light
   (c) It only works in flat spacetime
   (d) It neglects time dilation effects
10. The example of inertial frame in our everyday life is:
   (a) A car accelerating on a highway
   (b) A train moving smoothly at constant speed
   (c) A person standing on a spinning platform
   (d) an airplane encountering turbulence

Section (B): CRQs (Short Answered Questions):

1. Show that for values of v << c, Lorentz transformation reduces to the Galilean transformation.
2. If a particle could move with the velocity of light, how much K.E. would it possess?
3. Explain the difference between Special and General Relativity in simple terms.
4. Differentiate between Inertial Frames of Reference and Non- Inertial Frames of Reference.
5. Why can't any object move at the speed of light?
6. What is the limitation in the Galilean Transformation Equation, and how did Lorentz solve it?
7. Calculate the value of γ (Lorentz factor) if the object is moving at the speed of light.

Section (C): ERQs (Long Answered Questions):

1. State and explain the basic postulates of Einstein's special theory of relativity. Discuss length-contraction, mass variation and time-dilation.
2. Explain the concept of mass-energy equivalence. Derive Einstein's mass-energy relation and demonstrate that 1 atomic mass unit (u) is equivalent to 931 MeV.
3. Discuss the important conclusions derived from General Theory of relativity. What are the experimental observations in favour of these conclusions?
4. How does it principle of relativity differ from the classical Galilean view?
5. Explain the concept of spacetime curvature in general relativity and how it is used to visualize the effects of gravity.
6. Derive the basic equations of the Galilean transformation and explain how they relate the positions and velocities of objects in different inertial frames.
7. Discuss the Lorentz transformation equations in special relativity and how they describe the relationship between space and time coordinates in different inertial frames. Give examples to illustrate their application.

Section (D): Numerical:

1. A rod 1 meter long is moving along its length with a velocity 0.6c. Calculate its length as it appears to an observer (a) on the earth (b) moving with the rod itself.
   [Ans: (a) 0.8m, (b) 1m]
2. How fast would a rocket have to go relative to an observer for its length to be contracted to 99% of its length at rest? **[Ans: 42.45×10^6 m/s]**
3. A particle with a proper lifetime of 1μs moves through the laboratory at 2.7×10^8 m/s. (a) What is its lifetime, as measured by observers in the laboratory? (b) What will be the distance traversed by it before disintegrating? **[Ans: (a) 2.3×10^{-6}s, (b) 620m]**
4. At what speed is a particle moving if the mass is equal to three times its rest mass?
   [Ans: $\frac{2\sqrt{2}}{3}$c]

5. If 4 kg of a substance is fully converted into energy, how much energy is produced? **[Ans: 3.6×10^{17}J]**
6. Calculate the rest energy of an electron in joules and in electron volts. **[Ans: 8.2×10^{-14}J, 0.511MeV]**
7. Calculate the K.E. of an electron moving with a velocity of 0.98 times the velocity of light in the laboratory system. **[Ans: 3.3×10^{-13}J]**
8. At what velocity does the K.E. of a particle equal its rest energy? **[Ans: $\frac{\sqrt{3}}{2}c$]**
9. A particle of rest mass m_0 moves with speed $c/\sqrt{2}$. Calculate its mass, momentum, total energy and kinetic energy. **[Ans: $\sqrt{2}m_0$, m_0c, $\sqrt{2}\ m_0c^2$, $0.41\ m_0c^2$]**
10. The nearest star to Earth is Proximal Centauri, 4.3light- years away. A spaceship with a constant speed of 0.800c relative to the Earth travels toward the star.
    (a) How much time would elapse on a clock as measured by travelers on the spacecraft? **[Ans: (a) 3.22 years]**
    (b) How long does the trip take according to Earth observers? **[Ans: (b) 5.38 years]**

Introduction:

Quantum physics is the study of matter and energy at atomic and sub-atomic level. It aims to discover the properties and behaviors of the very building blocks of nature. At these scales, the classical laws of physics no longer apply, and strange, seemingly random phenomena govern the behavior of particles. In early 20th century, the field of quantum physics was developed as a result of certain experimental discoveries which could not be described by classical physics. The fundamental discovery of quantum physics was that energy is radiated as discrete packets, or quanta. Quantum physics explores the principles of wave-particle duality, uncertainty, superposition and quantization, which form the basis of our understanding of the behavior of matter, energy, and the fabric of reality itself.

25.1 Quantum Theory of Radiation:

It was introduced by Max Planck and Albert Einstein It explains how energy is emitted and absorbed in discrete packets (quanta), the key points of Quantum theory of radiation are:

- Energy is quantized (small packets, not continuous)
- Frequency (not amplitude) determines energy of radiation
- Radiation has both wave and particle properties (duality)

This theory revolutionized our understanding of light and radiation!

25.1.1 Black Body Radiation:

A black body is an idealized object which absorbs all incident radiation and also emits it in a continuous spectrum of colors, depending upon its temperature as illustrated in fig 25.1. This evidence encouraged physicists to explain the mechanisms through which entities like stars and heated metals produce both light and heat.

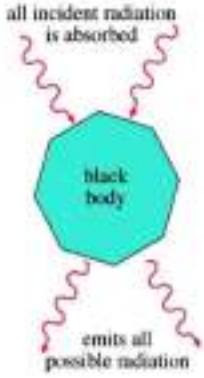


Figure 25.1
Black body Radiation

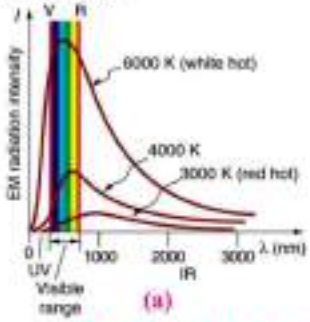


Figure 25.2 (a, b) Radiation intensity and red-hot horse

The spectrum of light emitted by heated objects was unexplained until 19th century. At normal temperature, we are not aware of this electromagnetic radiation due to low intensity. At higher temperatures, infrared radiation is emitted that we can feel heat if we are close to the object. At higher temperatures, objects actually glow, such as red-hot horse-shoe as shown in figure 25-2 (b). At temperatures more than 2000 K, objects glow with a yellow or whitish

color, such as white-hot iron. The light emitted contains a continuous range of wavelengths or frequencies, and the spectrum is a plot of intensity versus wavelength or frequency. As the temperature increases, the electromagnetic radiation emitted by objects not only increases in total energy but has its peak intensity and higher frequencies. Figure 25.2 Graphs of blackbody radiation (from an ideal radiator) at three different radiating temperatures.

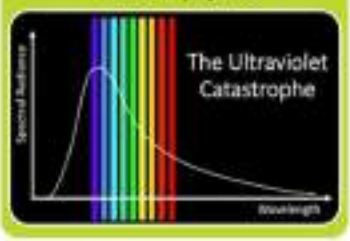
The intensity or rate of radiation emission increases directly with temperature, and the peak of the spectrum shifts toward the visible and ultraviolet parts of the EM spectrum. The shape of the spectrum cannot be described with classical physics.

Some experimental facts about blackbody radiations:

- The blackbody spectrum depends only on the temperature of the object, and not on what it is made of i.e., material. An iron horseshoe, a ceramic vase, and a piece of charcoal --- all emit the same blackbody spectrum if their temperatures are the same. As the temperature of an object increases, it emits more intense blackbody radiations of all wavelengths.
- As the temperature of an object increases, the peak wavelength of the blackbody spectrum curve shifts towards shorter wavelength (Blue-shift). For example, blue stars are hotter than red stars.
- The blackbody spectrum peak shifted always becomes small at the left-hand side i.e., the short wavelength, high frequency.

DO YOU KNOW?

"Ultraviolet catastrophe," where the energy output would become infinite as the wavelength approached zero. This contradicted experimental observations and posed a significant challenge to classical physics.



25.1.2 Classical explanation of black body radiation:

Rayleigh-Jeans Law:

This Law states that the energy per unit volume per unit wavelength of blackbody radiation is inversely proportional to the fourth power of the wavelength (λ), not directly proportional to the square of the wavelength.

The law is given by the formula:

$$E(\lambda, T) = \frac{2ckT}{\lambda^4}$$

Mathematically expressed as:

$$E(\lambda, T) = \frac{Constant}{\lambda^4} \quad \ldots\ldots 25.1$$

where:

- $E(\lambda,T)$ is the spectral radiance
- λ is the wavelength,

The failure of the Rayleigh-Jeans Law highlighted the need for a new theoretical framework, eventually leading to Max Planck's development of the quantum theory of radiation in 1900. Planck's work successfully resolved the ultraviolet catastrophe and laid the foundation for quantum physics.

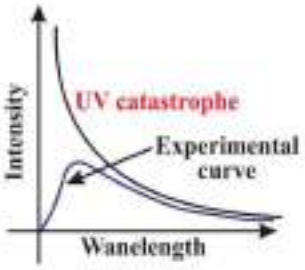


Figure 25.3 ultraviolet catastrophe

- The graph in figure depicts that for large

wavelengths it fitted the experimental data but it had major problems at shorter wavelengths.

Wien's Displacement Law:

This law breaks down completely at low frequencies. Wien's displacement law states that the wavelength at which the intensity of blackbody radiation is maximum is inversely proportional to the temperature of the blackbody. As the temperature increases, the peak of the blackbody radiation curve shifts to shorter wavelengths. This relationship is mathematically expressed as:

λ_{max} T=Constant

$$\lambda_{max} = \frac{Constant}{T} \ldots\ldots 25.2$$

Where λ_{max} is the wavelength at which a blackbody radiates at a given temperature T.

Stefan's Law:

The experimental relation is known as Stefan Boltzmann's law; it states that the total energy radiated per unit surface area of a black body is proportional to the fourth power of its temperature.

Mathematically, it is expressed as:

$P = \sigma AT^4$ …….. 25.3

Where A is the surface area of a blackbody, T is its temperature and σ is the Stefan's-Boltzmann constant,

$\sigma = 5.67 \times 10^{-8}$ W / (m^2. K^4).

DO YOU KNOW?

Stefan's law

It enables us to estimate how much energy a star is radiating by remotely measuring temperature.

25.1.3 Planck's Hypothesis:

According to Planck's Hypothesis, the energy levels of the oscillators in the walls of a blackbody are quantized, meaning they can only have specific, discrete values. This quantization of energy levels is what leads to the characteristic spectrum of blackbody radiation, which Planck's Law describes accurately.

Planck proposed that electromagnetic energy could only be emitted or absorbed in discrete packets, which he called "quanta." This assumption led to the concept of energy levels being quantized, meaning that the energy could only take on certain discrete values. The energy (E_n) of an oscillator in the cavity walls is given by the equation:

$$E_n = nhf \ldots\ldots\ldots\ldots 25.4$$

where: E_n is the energy of the oscillator is an integer (1, 2, 3, …), h is Planck's constant, f is the frequency of the oscillator.

This relationship, now known as Planck's Law, marked the birth of quantum theory, fundamentally changing our understanding of energy and leading to the development of quantum mechanics

An oscillator in the wall can receive energy from the radiation in the cavity (absorption), or it can emit energy to the radiation in the cavity (emission). The absorption process sends the oscillator to a higher quantum state, and the emission process sends the oscillator to a lower quantum state. Whichever way this exchange of energy goes, the smallest amount of energy allowed to exchange is ***hν***. There is no upper limit to how much energy can

be exchanged, but whatever is exchanged must be an integer multiple of ***hν***. If the energy packet does not have this exact amount, it is neither absorbed nor emitted at the wall of the blackbody.

Self-Assessment Questions:

1. How does the energy distribution of black body radiation change with temperature? Describe the relationship between temperature and the peak wavelength of radiation.
2. Describe the ultraviolet catastrophe problem in the context of black body radiation. How did Max Planck's quantum hypothesis resolve this issue?

Worked Example 25.1

At the Earth's surface, the energy flux in sunlight is 1.02×10^3 watt/m^2. If a black sheet of paper faces the Sun, what is the equilibrium temperature of the paper? Assume that the bottom of the paper is insulated, so the only heat loss is by blackbody radiation from the top surface.

Solution:

Step 1: Energy = P / A = 1.02×10^3 watt/m^2

T =?

$\sigma = 5.67\times10^{-8}$ watt/m^2. K^4

Step 2: Write down the formula and rearrange if necessary.

$E = \sigma T^4$

Step 2: Put the values in above formula $\sigma T^4 = 1.02\times10^3$ $T = \left[\frac{1.02\times10^3}{5.67\times10^{-8}}\right]^{1/4}$

T = 362 K or

T = 90 °C

Result: T = 90 °C

25.1.4 Particle Nature of Electromagnetic Radiation:

Electromagnetic radiation can be described in terms of a stream of **mass-less particles, called photons**. Each photon contains a certain amount of energy. The different types of radiation are defined by the amount of energy found in the photons.

The particle nature of light states that light consists of particles called 'Photons' (Particles do not interfere) That is, when the space is occupied by some particle, other particles cannot occupy the same space. Experiments from the last century, such as the **photoelectric effect** and **atomic line spectra**, can only be explained if EM radiation is assumed to behave as particles. The energy of a photon depends upon the frequency as described in Planck's hypothesis.

$$E = h\nu$$

$$But\ f = \frac{1}{t}\ and\ V = \lambda\nu$$

$$c = \lambda\nu$$

$$E = \frac{h\nu}{\lambda} \ldots\ldots\ldots 25.5$$

h = Planck's constant = h=6.626×10^{-34}

c = Speed of light = 3×10^{8} m/s

λ = Wavelength of Photon particle

In EM spectrum as you move right wards the wavelength decreases and frequency increases; consequently, energy increases.

Worked Example 25.2

Calculate the energy of a photon with a frequency of 3×10^{18} Hz. Analyze the question
Determine the energy of a photon given the frequency f= 3×10^{18} Hz. The frequency is in standard units and we know the relationship (E=hv) between frequency and energy.

Solution:

Step 1: given data

f= 3×10^{18} Hz

Energy =?

Step 2: Write down the formula and rearrange if necessary.

E=hv

Step 2: Put the values in above formula

E=hv

=6.63×10^{-34} J·s ×3×10^{18} Hz=2×10^{-15}J

Result: E= 2×10^{-15}J

Self-Assessment Questions:

1. Ina microwave oven photon of energy1.05×10^{-5} eV is used to heat food.
   (a) Calculate the frequency of the photons.
   (b) Calculate the wavelength of the photons.

25.2 Photoelectric Effect:

- In 1887, Heinrich Hertz observed the photoelectric effect in Germany, discovering that ultraviolet (UV) light shining onto metal surfaces causes the ejection of electrons, which are pushed with potential to reach an anode.
- In 1899, J.J. Thomson in England demonstrated that the incidence of UV light on a metal surface causes the ejection of electrons.
- In 1902, Philip Lenard measured the photoelectric effect quantitatively and concluded that light has electrical properties.

Though these experiments were able to describe phenomenon of photoelectric effect but some situations during experiments were not answered with scientific logic.

In 1905, Albert Einstein was able to resolve the phenomenon. Photoelectric effect is defined as:

When electromagnetic radiation like light is shined on certain metallic materials; electrons are emitted due to absorption of light by the electrons on the surface of material.

25.2.1 Photoelectric Effect Phenomenon:

The photoelectric phenomenon is shown in figure 25.4 in which an evacuated tube containing two electrodes cathode and anode; connected with a variable source voltage V. A monochromatic light source is shined onto cathode through a quartz window and anode (collecting electrode) is connected either positive or negative potential with respect to the cathode. An ammeter is connected with circuit to record the current due to photo-electrons.

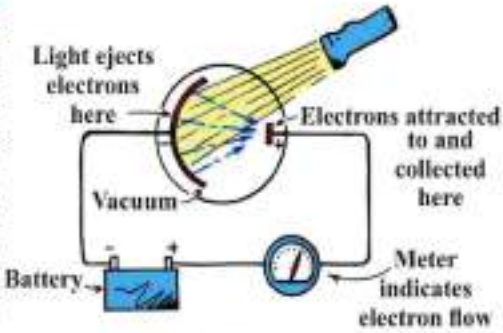


Figure 25.4 Photoelectric Effect

Following observations were made during this experiment:

1. For a constant potential difference between the cathode and anode, the number of electrons emitted from cathode increases with increasing intensity of radiation.
2. For a constant intensity and frequency of incident radiation the photoelectric current varies with the potential difference V between the cathode and anode and reaches a constant value beyond which further increase of potential difference does not affect the photoelectric current, instead, if the anode is made more and more negative with respect of the photo-cathode surface the current decreases. This negative potential (with respect to cathode) of the plate is called ***retarding potential***. As shown in fig: 25.5.
   For a particular value of retarding potential, the photoelectric current becomes zero.
   This potential is called ***cut-off or stopping potential*** V_0 and is measure of maximum kinetic energy of photo-electrons and we can write
   $$K.E_{max} = eV_0$$
   where $K.E_{max}$ (the maximum kinetic energy of the ejected electron.)

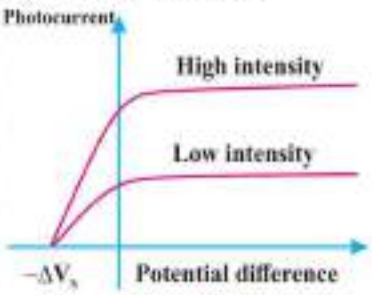


Figure 25.5
graph of photo current vs potential

DO YOU KNOW?

Einstein was awarded the 1921 Nobel prize in physics for his work on the photoelectric effect.

3. The stopping potential and hence the maximum kinetic energy $K.E_{max}$ of photo-electrons is independent of the intensity of incident radiation and depends only on the frequency ν of radiation.
4. For each substance there exists a characteristic frequency ν_0 such that for radiation with frequency below the photo-electrons are not ejected from the surface. This frequency is called the ***threshold frequency*** and the corresponding wavelength is called ***Threshold wavelength***, λ_0 $(\frac{hc}{\varphi})$
5. The time lag between the incidence of radiation and the emission of a photoelectron is very small, less than 10^{-9}seconds.

25.2.2 Explanation of photo-electric effect on Quantum Theory:

Einstein proposed that radiation energy is not continuously distributed over the wave-front, but the light energy consists of discrete quanta of energy $h\nu$,which penetrates the surface of the cathode, all of its energy is transformed to an electron; it's depicted in figure 25.6. The photon's energy would be associated with its frequency ν, through a proportionality constant known as Planck's constant h, or alternately, using the wavelength λ and the speed of light c:

$$K.E = h\nu = \frac{hc}{\lambda}$$

or the momentum equation:

$$p = \frac{h}{\lambda}$$

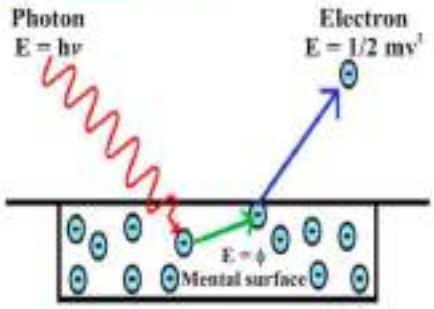


Figure 25.6 photo-electric effect

When a photon is incident on the surface of a material, some of its energy is spent in making the electron free and the rest appears as kinetic energy of the electron. The electrons at the surface of the material are most loosely bound and require minimum energy for their liberation. This energy is called the ***work function*** ϕ of the material. The maximum kinetic energy of photo electrons ejected from the surface is given by

$$K.Emax = \frac{1}{2}mv^2 = h\nu - \phi \quad \text{..... 25.6}$$

So, ***work function is the minimum energy that must be supplied to the electron for it to leave the Metal surface. As shown in fig: 25.7.***

Table 25.1
Threshold frequencies value

Metal	Work function ϕ	
Name	*Symbol*	*eV*
Sodium	Na	2.4
Calcium	Ca	2.8
Potassium	K	2.3
Copper	Cu	4.4
Silver	Ag	4.3
Lead	Pb	4.0
Platinum	Pt	6.4
Aluminum	Al	4.28
Gold	Au	5.10
Iron	Fe	4.7
Tungsten	W	4.55

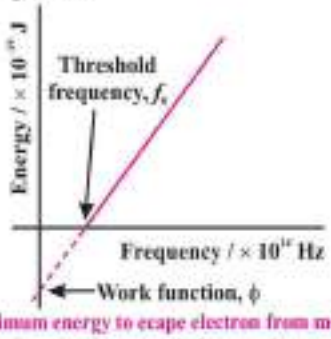


Figure 25.7 a minimum energy to ecape electron from metal surface graph

Consequently there exists a minimum frequency that is independent of the intensity of light, below which electrons cannot be ejected from the metal. This proposal was proved by Millikan when in 1914 he published his results of the voltage required to stop photoelectrons ejecting from a metal surface.

Self-Assessment Questions:

1. How does the photoelectric effect demonstrate the particle nature of light?
2. Calculate the threshold frequency of incident radiation that will cause the emission of photo electrons from the surface of gold.

Worked Example 25.3

The work function for zinc is 4.24 eV. What is the threshold frequency for the ejection of photoelectrons from zinc?

Step 1: When an electron absorbs a photon at threshold frequency, it will just barely have enough energy to overcome the binding forces holding it in the metal and it will emerge with zero kinetic energy.

Step 2: Write down the formula and rearrange if necessary.

$$K.Emax = \frac{1}{2}mV^2 = h\nu - \phi$$

Step 3: First, we determine the frequency in terms of the wavelength.

$$K.Emax = \frac{1}{2}mV^2 = h\nu - \phi$$

$$0 = h\nu - \varphi_o$$

$$\nu = \frac{\phi}{h} = \frac{4.24 \times 1.6 \times 10^{-19}}{6.63 \times 10^{-34}}$$

$$\nu = 1.02 \times 10^{15}\ Hz$$

Result: $\nu = 1.02 \times 10^{15}\ Hz$

Photoelectric effect in solar cells and photocells:

Photovoltaic cells are made up of p-type and n-type silicon semiconductors. When light falls on the junction between p-type and n-type semiconductors, electrons are emitted according to the photoelectric effect. These electrons are collected in an external circuit to produce an electric current.

- Photocell, a device that uses the photoelectric effect to generate or control a current.
- Photovoltaic cells and photoconductive cells are examples of photoelectric cells.
- Solar cell, a device that converts electromagnetic energy into electrical energy.
- Photovoltaic cells are used to power a wide variety of devices, and are often used in conjunction with a battery that stores the power they produce.

25.3 Compton Effect:

A. H. Compton (1892-1962), aimed short-wavelength light (i.e., x-rays or γ-rays) at various materials, and detected light scattered at various angles. He observed that the scattered light had a slightly longer wavelength than did the incident light, and therefore a slightly lower frequency indicating a loss of energy. He explained this result based on the photon theory as incident photons colliding with electrons.

25.3.1 Particle Model of Photons with Energy and Frequency:

In Compton Effect experiment, there is increase in wavelength of photons, due to their scattering by an electron. The impact results in one of the fundamentals of quantum mechanics, which represents wave and particle properties of light.

Arthur Holly Compton successfully demonstrated that X-rays can be treated as discrete bundles, or quanta, of electromagnetic energy. This concept was later termed "photon" by American physicist Gilbert Lewis.

Photons exhibit both particle-like properties, such as energy and momentum, and wave-like properties, such as frequency and diffraction. The energy of a photon is dependent on its frequency, with lower energy photons having lower frequencies and longer wavelengths, as described by the equation $E = \frac{h\nu}{\lambda}$.

In Compton's experiment, photons collide with free or loosely bound electrons in matter. During these collisions, photons transfer part of their energy and momentum to the electrons, causing the electrons to recoil. This interaction results in the emission of new photons with lower energy and longer wavelengths, a phenomenon known as the Compton shift. The shift in frequency of the scattered photons depends on the energy transferred to the electrons and is independent of the initial wavelength of the incident photons.

25.3.2 Compton Effect Qualitatively:

Compton's Effect, as illustrated in Figure 25.8, can be qualitatively explained as follows:

Initial Interaction:

- A photon, which is a packet of electromagnetic energy, interacts with an electron in a material, typically a target like a metal or graphite.

Scattering Process:

- During the interaction, the photon transfers some of its energy and momentum to the electron. This transfer causes the photon to change its direction and wavelength (or equivalently, its frequency).

Change in Photon Energy:

- The scattered photon emerges with less energy (longer wavelength) than the initial incident photon. The amount of energy lost by the photon is directly related to the energy and momentum gained by the electron.

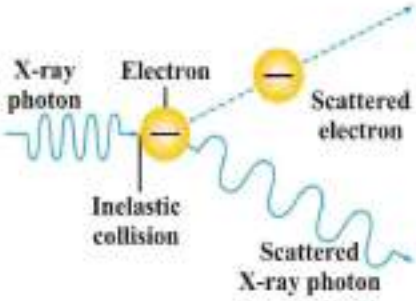


Figure 25.8 Compton Effect

Quantum Nature:

- Compton's Effect cannot be explained using classical wave theory alone. Classical wave theory predicts that light should scatter uniformly without a change in wavelength. However, Compton's observations demonstrated that the scattered light has a shifted wavelength, indicating a particle-like interaction.

Experimental Confirmation:

- Compton conducted experiments where X-rays were targeted at graphite and the scattered X-rays were observed. By measuring the scattering angle and change in wavelength, he confirmed that the results were consistent with the predictions of quantum theory.

Wavelength Shift:

- The wavelength shift observed in Compton scattering is directly proportional to the Compton wavelength, which depends on the mass of the electron and the Planck constant. This relationship provides crucial evidence for the particle nature of photons.

The Compton shift can be evolved as under:

$$\lambda_2 - \lambda_1 = \frac{hc}{E_o}(1 - Cos\theta) = \frac{h}{m_0 c}(1 - Cos\theta)$$

$$\Delta\lambda = \frac{h}{m_0 c}(1 - Cos\theta) \quad \ldots\ldots\ldots\ldots 25.7$$

This equation describes the phenomenon known as Compton Effect. $\Delta\lambda$gives the change in photon wavelength due to scattering with a free electron and it is called **Compton shift.**

It's clear that the Compton Shift is independent of the wavelength of the incident photon and depends on scattering angle.

The term $\lambda_c = \frac{h}{m_0 c} = 2.426 \times 10^{-12} m$, is called Compton Wavelength of the scattering particle i.e., electron.

Self-Assessment Questions:

1. Compare and contrast the photoelectric effect with the Compton effect and the Bohr model of the hydrogen atom in terms of their contributions to the development of quantum theory.
2. How does the Compton Effect demonstrate the particle nature of light?

Worked Example 25.4

A beam of x-rays of wavelength 0.01 Å is incident on a block of graphite. What is the wavelength of the x-rays scattered at an angle of 30°?

Step 1: given data and what to be found

$\theta = 30°$

$\lambda = 0.01$ Å

Step 2: Write down the formula and rearrange if necessary.

$$\Delta\lambda = \frac{h}{m_0 c}(1 - Cos\theta)$$

Step 3: $\Delta\lambda = 2.426 \times 10^{-12}(1 - Cos\, 30°) = 0.003$ Å

Result: Hence, the final wavelength of these x-rays is $\lambda + \Delta\lambda = 0.01 + 0.003 = 0.013$ Å

25.5.1 Conservation laws of annihilation of matter:

$$e^- + e^+ \rightarrow \gamma + \gamma$$

Each photon has an energy equal to the rest mass of electrons 0.51 MeV. The two photons are produced moving in opposite direction in order to conserve momentum and energy which forbids the creation of a single gamma- ray, as shown in figure 25.10. The charge is also conserved as net charge before and after is zero.

Pair production and Annihilation of matter supports that energy and mass are inter convertible i.e. $E=mc^2$.

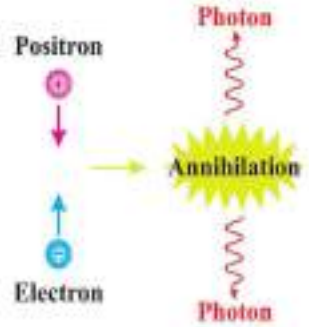


Figure 25.10 Annihilation of Matter

25.6 Wave Nature of Particles:

Classical physics traditionally treated particles and waves as separate and distinct entities.

However, the advent of quantum theory proposed a dual character for radiations, suggesting that they could exhibit characteristics of both waves and particles.

The wave nature of particles implies that particles can behave like waves, displaying properties such as reflection, refraction, interference, diffraction, and other wave-like characteristics. In essence, this duality suggests that matter and radiation can co-exist both particle and wave attributes.

Examples of particles that exhibit this dual nature include:

(i) Matter particles, such as electrons, protons, and neutrons.
(ii) X-rays.
(iii) Photons.
(iv) Electromagnetic radiation.

Particles, like above, can behave like waves. This means they can show wave-like properties, like diffraction and interference. This dual behavior (both wave-like and particle-like) is a fundamental principle of quantum mechanics, which helps us understand the behavior of matter and radiation at a tiny scale.

25.6.1 De Broglie wave:

Louis de Broglie's proposal of wave-particle duality for particles had a profound impact on the understanding of the nature of matter and laid the foundation for the development of quantum mechanics.

De Broglie's idea suggested that both matter and energy could exhibit both particle and wave properties. This unified description was a departure from the classical distinction between particles and waves, providing a more comprehensive understanding of the behavior of particles at the quantum level.

De Broglie's proposal was instrumental in the development of Niels Bohr's model of the atom. It helped explain why electrons in quantized orbits do not radiate energy continuously but exists in stable orbits, as their orbits are analogous to standing waves.

De Broglie Wavelength:

According to de Broglie hypothesis, ***all matter particles like electrons, protons and neutrons in motion are associated with waves.*** These waves are called de Broglie waves or matter waves. The momentum of photon of frequency f is given by

$$p = \frac{hf}{c} = \frac{h}{\lambda} \qquad \text{since } c = v\lambda$$

The wavelength of a photon in terms of its momentum is

$$\lambda = \frac{h}{p} \qquad \text{but } p = mv, \text{ then}$$

According to de Broglie, the above equation is completely applicable to matter particles as well. Therefore, for a particle of mass m travelling with speed v, the wavelength is given by

$$\lambda = \frac{h}{mv} \quad \ldots\ldots\ldots\ldots \quad 25.8$$

This wavelength of the matter waves is known as de Broglie wavelength. This equation relates the wave character (the ***wavelength***) and the particle character (the ***momentum***) through 25-3, Planck's constant.

Self-Assessment Questions:

1. Calculate the de Broglie wavelength of an electron travelling through a space at a speed of 10^7 m/s. State whether or not these electrons be diffracted by solid materials (atomic spacing in solid materials is 10^{-10} m).
2. How does the De Broglie wavelength relate to the momentum of a particle?

25.6.2 Davisson – Germer Experiment:

De Broglie hypothesis of matter waves was experimentally confirmed by Clinton Davisson and Lester Germer in 1927. They demonstrated that electron beams are diffracted when they fall on crystalline solids. Since crystal can act as a three-dimensional diffraction grating for matter waves, the electron waves incident on crystals are diffracted off in certain specific directions. Figure 25.11 shows a schematic representation of the apparatus for the experiment.

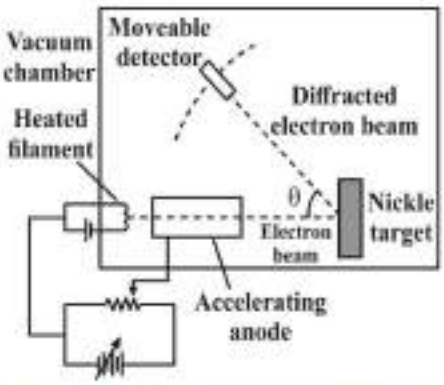


Figure 25.11 Davisson Germer Experiment

The experimental parts are given as under:

1. An electron gun emits electrons via thermionic emission produced by the tungsten filament used in it, i.e., when heated to a specific temperature.
2. Two opposite charged plates known as Electrostatic Particle Accelerator, which accelerates the electrons at a certain potential.

3. The accelerator is enclosed within a cylinder called Collimator which is a narrow passage for the electrons along its axis.
4. The target is a Nickel crystal on which the electron beam is fired normally.
5. When the electrons are scattered from Ni crystal, these are captured by the semicircular movable detector.

Davisson Germer Experiment and de Broglie Relation:

Let us consider an electron of mass (m_0), charge (e) accelerated from rest through a potential V. Then, the kinetic energy $K.E$ of the electron equals to the work done (eV) on it by the electric field:

$K.E = eV$

Now, $K.E = \frac{1}{2}m_0v^2 = \frac{p^2}{2m_0}$

$$p = \sqrt{2m_0K.E} = \sqrt{2m_0eV}$$

Then, the de Broglie wavelength of electron is given by

$\lambda = \frac{h}{p} = \frac{h}{\sqrt{2m_0eV}}$ 25.9

By using $h = 6.626\times10^{-34}$ J-s, $m_0 = 9.11\times10^{-31}$ kg, and $e=1.6\times10^{-19}$ C

we get

$$\lambda = \frac{1.227}{\sqrt{V}} \text{ nm}$$

Where V is the magnitude of accelerating potential in volts.

As, de Broglie wavelength is associated with electrons for V = 54 V from graph 25.12., then we have

$\lambda = 0.167$ nm

DO YOU KNOW?

Bragg's Equation

Electron Diffraction:

- Known: crystal spacing (d) from X-ray diffraction
- Measured: angles (θ) for diffraction peaks
- Calculated: wavelength (λ) of matter waves using: $n\lambda = 2d\sin(90^\circ - \theta/2)$
- Result: $\lambda = 0.165$ nm (wavelength of electron matter waves)

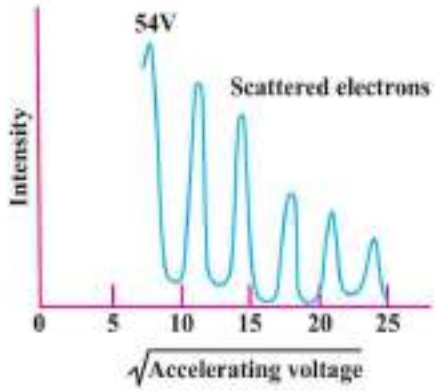


Figure 25 .12 de Broglie wavelength is associated with electrons graph

We conclude that this experiment confirms the wave nature of electrons and the de Broglie relation.

25.6.3 Electron Microscope:

De Broglie hypothesis are used in the field of electron optics. In the electron microscope design, the wave properties of electrons have been utilized with higher resolution, which is giving a great improvement in visualization.

A special type of microscope that uses an electron beam with its wavelike properties to illuminate a specimen able to magnify objects in high resolution (nanometers), which are formed by controlled use of electrons in a vacuum captured on a phosphorescent screen. **Ernst Ruska** (1906-1988), a German engineer and academic professor, built the first Electron Microscope (EM) in 1931, and the same principles behind his prototype still govern modern EMs.

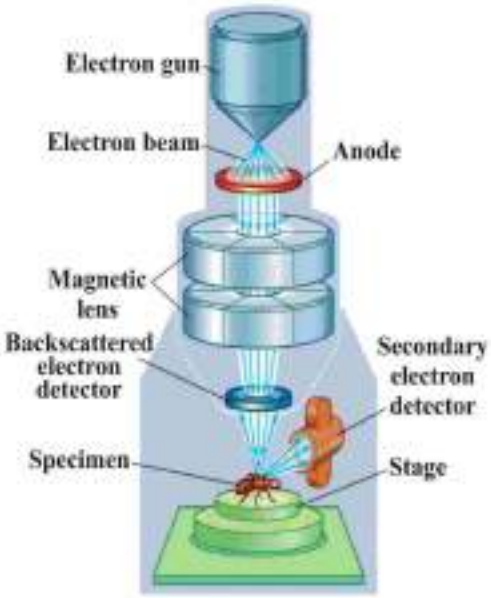


Figure 25.13 Electron Microscope

De Broglie hypothesis of electron have wave paved the way for the development of the electron microscope (EM), which can produce images of much greater magnification than an optical microscope. There are two types of electron microscopes which are transmission electron microscope (TEM), which produces a two-dimensional image, and the scanning electron microscope (SEM), which produces images with a three-dimensional.

Working:

In both types i.e., TEM & SEM, the objective and eyepiece lenses are actually magnetic fields that exert force on the electrons to bring them to a focus. The fields are produced by carefully designed current-carrying coils of wire. EM's measure the intensity of electrons, producing mono chromatic image. Color is often added artificially to highlight. The probe tip of scanning electron microscope, as it is moved horizontally, automatically moves up and down to maintain a constant tunneling current, and this motion is translated into an image of the surface.

Applications:

- Electron microscopes are used to investigate the ultra-structure of a wide range of biological and inorganic specimens including microorganisms, cells, large molecules, biopsy samples, metals, and crystals.
- Industrially, electron microscopes are often used for quality control and failure analysis.

DO YOU KNOW?

Electron microscopes
As the wavelength of an electron can be up to 100,000 times shorter than that of visible light photons, electron microscopes have a higher resolving power than light microscopes and can reveal the structure of smaller objects.

- Modern electron microscopes produce electron micrographs using specialized digital cameras and frame grabbers to capture the images.
- The advancement of microbiology is significantly indebted to the electron microscope, which has revolutionized our understanding of microorganisms such as bacteria, viruses, and other pathogens, thereby greatly enhancing the effectiveness of disease treatments.

25.6.4 Role of Electron Microscope:

1. **Higher Resolution:** Electron microscope offers significantly higher resolution compared to optical microscopes, allowing observation at the nanometer scale due to the shorter wavelength of electrons.
2. **Transmission Electron Microscope (TEM):** TEMs study internal structures of thin specimens, producing detailed images of cells, organelles, and crystalline structures. Valuable in biology, materials science, and nanotechnology.
3. **Scanning Electron Microscope (SEM):** SEMs provide 3D surface images by scanning specimens and detecting emitted secondary electrons. Widely used in biology, geology, and materials science for surface analysis.
4. **Energy- Dispersive X-ray Spectroscopy (EDS):** With EDS detectors, electron microscopes analyze elemental composition by detecting X-rays emitted when high-energy electrons interact with the sample.
5. **Materials Science:** Crucial for studying the microstructure of materials, aiding understanding of relationships between microstructure and properties in metals, ceramics, and polymers.
6. **Nanotechnology:** Essential for imaging and characterizing nano-materials, contributing to the development of new materials and devices through observation and manipulation of nano-particles and nanostructures.

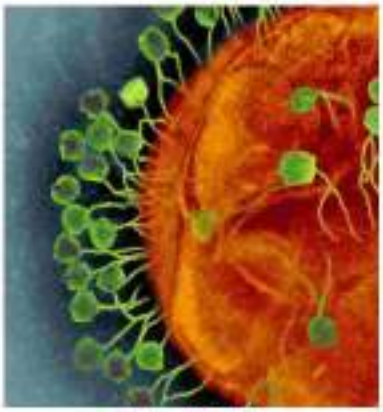


Figure 25.14 (a)
Viruses attacking a cell of bacterium (×50000) by TEM

Figure 25.14 (b)
Viruses attacking a cell of bacterium (×35000) by SEM

7. **Advancements in Medicine:** Contributes to medical research by providing insights into the structure of viruses, bacteria, and cellular organelles, aiding understanding of diseases and therapeutic development.
8. **Quality Control in Industry:** Used across industries for quality control and failure analysis, helping identify defects, analyze material composition, and ensure product integrity at the microscopic level.

25.7 Uncertainty Principle:

In 1927, Werner Heisenberg suggested a principle that applies to measure the properties of quantum-sized objects (e.g., atomic and sub-atomic particles).

The uncertainty principle mainly applied to experiments in physics labs, however some real-world effects are given. For example, in the sun's core, it is the uncertainty in the position of two hydrogen nuclei (protons) that allows for there to be a chance that they will overlap and fuse together. This fusing releases a tremendous amount of energy, which we receive as light. A pretty significant effect on our everyday lives.

25.7.1 Heisenberg Uncertainty Principle:

It states that it is impossible to measure simultaneously, certain pairs of properties of a subatomic particle, such as its position and momentum, with absolute certainty.

In other words the more precisely we know the position of a particle, the less precisely be able to know its momentum, and vice versa.

This principle develops from the wave-particle duality of particles in quantum physics, where particles can exhibit both wave-like and particle-like behaviors.

The formal inequality relating the standard deviation of position Δx and the standard deviation of momentum Δp are given here:

$$\Delta x . \Delta p \geq \frac{h}{4\pi}$$

$\Delta x, \Delta p \geq \frac{\hbar}{2}$.....25.10

ħ (pronounced "ħ-bar") is the reduced Planck constant

Heisenberg uncertainty principle can also be applied to other pairs of complementary quantum properties, such as energy and time

$\Delta E . \Delta t \geq h/2\pi$ 25.11

Here, ΔE and Δt represents the uncertainties in the energy and time respectively.

The Uncertainty Principle has profound consequences for understanding quantum world. It means that at the subatomic level, there are inherent limits to how precisely we can measure certain properties of particles. This is not due to restrictions in our measuring instruments but it is a fundamental aspect of the quantum nature of the universe.

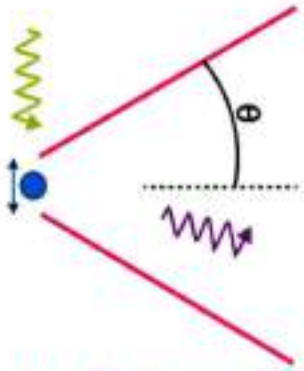


Figure 25.16 Uncertainty Principle

DO YOU KNOW?

Interesting Facts about Planck's Constant (h):

Universal constant: Appears in many areas of physics, like quantum mechanics, thermodynamics, and electromagnetism.

Tiny but mighty: Very small value (approximately 6.626 x 10^-34 J s), but has a huge impact on our understanding of the universe.

Fundamental unit: Used to define the Planck units, a set of fundamental units for length, time, mass, and energy.

SUMMARY

- A black body absorbs all incident radiation and releases it in a continuous spectrum of colors, depending upon its temperature.
- The energy per unit volume per unit wavelength of black body radiation is directly proportional to the square of the wavelength and increases indefinitely as the wavelength becomes smaller.
- Different atoms and molecules can emit or absorb energy in discrete quantities only.
- When electromagnetic radiation like light is shinned on certain metallic materials; electrons are emitted due to absorption of light by the electrons on the surface of material.
- Work function, is the minimum energy that must be supplied to the electron for it to leave the metal.
- For each substance there exists a characteristic frequency f_0 such that for radiation with frequency below the photo-electrons are not ejected from the surface.
- Compton scattering, refers to X-ray or gamma-ray photons undergo a change in wavelength and momentum when they collide with charged particles, typically electrons.
- Change in wavelength gives the change in photon wavelength due to scattering with a free electron.
- When the photon interacts with the strong electric field around the nucleus, it undergoes a change of state and is transformed into two particles (creating matter from energy):
- Electron-positron annihilation occurs when an electron and a positron (the antiparticle of the electron) collide.
- All matter particles like electrons, protons and neutrons in motion are associated with waves.
- A microscope which uses a beam of accelerated electrons as a source of lighting.
- It is impossible to measure simultaneously all quantum properties of a particle with equal accuracy.

5. Of the following, Compton scattering from electrons is most easily observed for:
   (a) infrared light (b) visible light
   (c) ultraviolet light (d) X-rays
6. In the Compton scattering from stationary electrons the largest change in wavelength occurs when the photon is scattered through
   (a) 0° (b) 45° (c) 90° (d) 180°
7. A free electron has a momentum of 5.0×10^{-24} kg · m/s. The wavelength of its wave function is:
   (a) 1.3×10^{-8}m (b) 1.3×10^{-10} m
   (c) 2.1×10^{-11} m (d) 2.1×10^{-13} m
8. In Photoelectric Effect, when light fall on the surface of metal, the material should emit
   (a) Electrons (b) Protons
   (c) Positrons (d) Neutrons
9. Threshold frequency is defined as the ..... frequency of incident light which can cause photo electric emission
   (a) Maximum (b) Minimum
   (c) Average (d) highest
10. The amount of energy which is necessary to start photo electric emission is called:
   (a) Maximum (b) Average
   (c) Minimum (d) Littlest

Section (B): CRQs (Short Answered Questions):

1. Differentiate between wave and particle.
2. Is it possible for the de Broglie wavelength of a particle?
3. Estimated Broglie wavelength of a cricket ball on the pitch?
4. Differentiate between the continuous and discrete emission of radiation?
5. What is threshold frequency?
6. How has the photoelectric effect been applied in real-world technologies or devices, and what are its practical implications?
7. What are the advantages of an electron microscope over an optical microscope?
8. Is it possible to create only an electron through matter and photon interaction?
9. Give construction of electron microscope?
10. Elaborate the particle nature of electromagnetic radiation.

Section (C): ERQs (Long Answered Questions):

1. Describe how energy is distributed over the wavelength range for different temperatures.
2. Explain the particle model of light in terms of photons with particular energy and frequency.
3. Describe conservation laws in pair production and annihilation of matter.
4. Describe Compton's effect qualitatively.
5. Explain how the very short wavelength of electrons, and the ability to use electrons and magnetic fields to focus them, allows electron microscope to achieve very high resolution.
6. Describe the impact of de Broglie proposal that any kind of particle has both wave and particle properties.
7. Describe the confirmation of de Broglie proposal by Davisson and Germer experiment in which the diffraction of electrons by the surface layers of a crystal lattice was observed.

Section (D): Numerical:

1. The Sun's surface temperature is 5700 K.
   (i) How much power is radiated by the Sun?
   (ii) Given that the distance to Earth is about 200 Sun radii, what is the maximum power possible from one square kilometer solar energy installation?
   (iii) What is the wavelength of maximum intensity of solar radiation?
   (5.98×10^{7} W/m^{2}, 3.6×10^{26} Watts, 5.1×10^{-7} m)
2. The temperature of your skin is approximately 32 °C. What is the wavelength at which the peak occurs in the radiation emitted from your skin? **(9.05×10^{-5} m)**
3. An FM radio transmitter has a power output of 100 kW and operates at a frequency of 94 MHz, How many photons per second does the transmitter emit?
   (6.23×10^{-26} J, 1.61×10^{30} s^{-1})
4. A light source of wavelength illuminates a metal and ejects photoelectrons with a maximum kinetic energy of 1.0 *eV*. A second light source with half the wavelength of the first ejects photoelectrons with a maximum kinetic energy of 4.0 eV. Determine the work function of the metal. **(2 eV)**
5. A 430 nm violet light is an incident on a calcium photo electrode with a work function of 2.71 eV. Find the energy of the incident photons and the maximum kinetic energy of ejected electrons. **(2.88 eV, 0.17 eV)**
6. Cut-off frequency for the photoelectric effect in some materials is 8×10^{13}Hz. When the incident light has a frequency of 1.2×10^{14} Hz, the stopping potential is measured as – 0.16 V. Estimate a value of Planck's constant from these data and determine the percentage error of your estimation. **(5.17 eV, 1.30 V)**

7. The work function of some metals is listed below. The number of metals which will show photoelectric effect when light of 300 nm wavelength falls on the metal is:

Metal	Li	Na	K	Mg	Cu	Ag	Fe	Pt	W
φ in eV	2.4	2.3	2.2	3.7	4.8	4.3	4.7	6.3	4.75

(4 Metals)

8. X-rays with an energy of 300 keV undergo Compton scattering with a target. If the scattered X-rays are detected at 30° relative to the incident X-rays, determine the Compton shift at this angle, the energy of the scattered X-rays, and the energy of the recoiling electron. **(0.35 pm, 278keV, 22keV)**

9. A photon with a wavelength of 6.0×10^{-12} m collides with an electron. After the collision the photon wavelength is found to have been changed by exactly one (Compton Wavelength is 2.43×10^{-12} m).
   (i) What is the photon's wavelength after collision?
   (ii) Through what angle has been deflected in this collision?
   (iii) What is the angle for the electron after the collision?
   (iv) What is the electron's kinetic energy, in eV, after collision?

   (8.4×10^{-12} m, 90°, $< 90^0$, 5.9×10^4 eV)

10. Find the de Broglie wavelength of an electron in the ground state of hydrogen.

    (3.324°A)

11. Determine the minimum uncertainties in the positions of the following objects if their speeds are known with a precision of 1.0×10^{-5}m/s: (a) an electron and (b) a bowling ball of mass 6.0 kg. **(5.8cm, 33 m)**

$\nu = \frac{(E_n - E_p)}{h}$ using equation (6)

$$\nu = \frac{me^4}{8\varepsilon_0{}^2h^3}\left(\frac{1}{p^2} - \frac{1}{n^2}\right)$$

To express above equation in terms of wavelength, it is convenient to use

$c = \nu\lambda$

$$\frac{1}{\lambda} = \frac{me^4}{8c\varepsilon_0{}^2h^3}\left(\frac{1}{p^2} - \frac{1}{n^2}\right)$$

where $R_H = \frac{me^4}{8c\varepsilon_0{}^2h^3} = 1.097 \times 10^7 m^{-1}$, is called Rydberg constant.

$$\frac{1}{\lambda} = R_H\left(\frac{1}{p^2} - \frac{1}{n^2}\right) \qquad (26.8)$$

The possible series associated the different wavelengths can be explained from equation (8). The emission and absorption lines for all possible jumps up from the $n = 1$ level and down to the $n = 1$ level are said to be in the Lyman series. Similarly, other series are illustrated in the figure 26.6.

Worked Example 26.1

The electron in a hydrogen atom makes a transition from the n =3 energy level to the ground level (n = 1). Find the wavelength and frequency of the emitted photon.

Solution:

Step 1: Write down the known quantities and quantities to be found.

n =3

p = 1

$$R_H = \frac{me^4}{8c\varepsilon_0{}^2h^3} = 1.097 \times 10^7 m^{-1}$$

Step 2: Write down the formula and rearrange if necessary.

$$\frac{1}{\lambda} = R_H\left(\frac{1}{p^2} - \frac{1}{n^2}\right)$$

$$c = \nu\lambda$$

Step 3: Put the values in formula and calculate.

$$\frac{1}{\lambda} = R_H\left(\frac{1}{1^2} - \frac{1}{3^2}\right) = \frac{8}{9}R_H$$

$$\frac{1}{\lambda} = 0.975 \times 10^7 m^{-1}$$

$$\lambda = 102.56\ nm$$

$$\nu = \frac{c}{\lambda} = \frac{3.0 \times 10^8 m/s}{102.56 \times 10^{-9} m}$$

$$\nu = 2.925 \times 10^{15} Hz$$

26.4.2 He – Ne Gas Laser:

The helium-neon laser was the first continuous wave (CW) laser ever constructed. It was built in 1961 by **Ali Javan**, **Bennett**, and **Herriott** at Bell Telephone Laboratories. He-Ne lasers are commonly used in school laboratories and in older barcode readers. The schematic diagram of He-Ne laser is shown in figure 26.15.

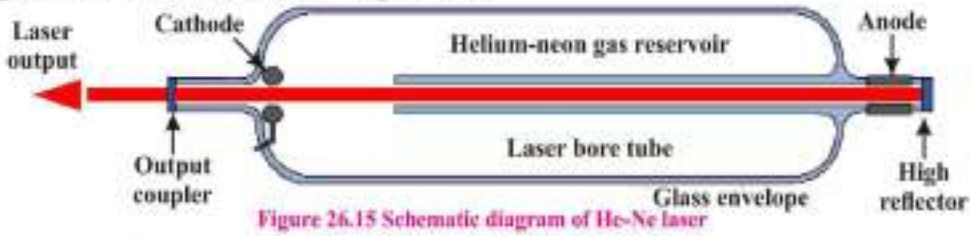


Figure 26.15 Schematic diagram of He-Ne laser

Helium-neon laser construction:

The helium-neon laser consists of three essential components:

Pump Source (high voltage power supply):

The pump energy of the laser is provided by an electrical discharge of several hundred Volts between an anode and cathode at each end of the glass tube. A current of 5 to 100 mA is typical for laser operation.

Gain Medium (laser glass tube or discharge glass tube):

Figure 26.15 shows a gas discharge tube contains a low-pressure mixture of helium-neon in a ratio between 5:1 and 20:1 bound in a glass tube. The partial pressure of helium is 1 mbar whereas that of neon is 0.1 mbar.

Resonating Cavity:

The glass tube (containing a mixture of helium and neon gas) is placed between two parallel mirrors. These two mirrors are silvered or optically coated. Each mirror is silvered differently. The left side mirror is partially silvered and is known as output coupler whereas the right side mirror is fully silvered and is known as the high reflector or fully reflecting mirror. The fully silvered mirror will completely reflect the light whereas the partially silvered mirror will reflect most part of the light but allows some part of the light to produce the laser beam.

DO YOU KNOW?

The three important components of any laser device are:

1. Active medium (medium in which the population inversion can be achieved)
2. Pumping source (usually light or electric current to achieve population inversion)
3. 3. Optical resonator (a feedback system, which consists of an active medium kept in between a 100% mirror and a partial mirror)

Laser Operation:

This electrical excitation raises the helium atom into a meta-stable state with energy 20.61 eV above the ground state. Neon has a meta-stable state with energy 20.66 eV above its ground state. Collision of the excited helium atoms with the ground-state neon atoms results in transfer of energy to the neon atoms, exciting neon electrons. The difference between the energy states of the two atoms is in the order of 0.05eV, which is supplied by kinetic energy. The number of neon atoms in the excited states builds up as further collisions between helium and neon atoms occur, causing a population inversion. Spontaneous and stimulated emission results in emission of 632.82 nm wavelength light.

26.4.3 Properties and Uses of Lasers:

Properties of Laser:

The properties of laser light are very strange as compared to conventional light. The factors that make laser light prominent are: Monochromatic, Highly Intense or Brightness Coherence and Directionality

1. Monochromatic:

A single color or wavelength light is called monochromatic. The light emitted from ordinary light sources has different wavelengths, or colors. However, laser light has a single wavelength or color.

2. Highly Intense or Brightness:

The characteristics of coherence and directionality of the laser makes laser light highly intense as compared to conventional light. It is the power emitted per unit surface area per unit solid angle. A one milli watt He-Ne laser is more intense than the sun intensity.

3. Coherence:

Two or more waves of same frequency are said to be coherent in nature if they have constant phase difference. A predictable correlation of the amplitude and phase at any one point with another point is called coherence. In laser light the property of coherence occurs between any two or more light waves. Where as in conventional light, the property of coherence exhibits between a source and its virtual source.

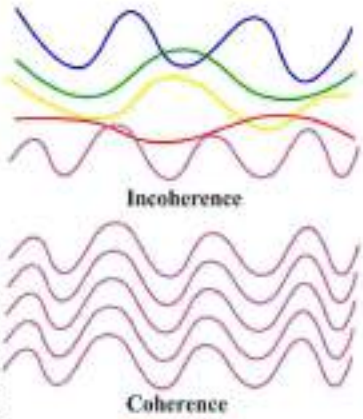


Figure 26.16:
Incoherence and Coherence

4. Directionality or Divergence:

The light ray coming from laser light travels in a single direction. However, an ordinary light source travels in all directions. For example, on travelling a distance of 1 Km the torch light spreads 1 km distance. But the laser light spreads only few centimeters distance even it travels lacks kilometer distance.

Uses of Lasers:

Lasers are widely used in many fields some of them are:

1. Tools:

- Cutting tools are typically precise, and also simple to automate. They do not require sharpening, unlike that of knives.
- Robot-guided lasers are widely used for the cutting of pieces of cloth.
- Laser tools are used in corneal surgery, restoring a detached retina of eye, and removing kidney stones.

2. Communication:

- Lasers are used in Barcode scanners to convert a printed barcode into a number.
- A semiconductor laser beam helps to convert the printed pattern of data into numbers in a CD/DVD.
- In Photonics, lasers are used in fibre optic cables.

3. Defence:

- In defence field, the military employs laser-guided guns and missiles.

- Laser range-finders use high-resolution scanning to find the distance and speed from an object that is located beyond the point-blank range.

4. Medicine:

- In medicine, the laser beam is used as part of phototherapy for many procedures.
- The development of laser technology in recent decades has enabled the creation of a new field of medicine laser surgery.
- Lasers have uses in dermatology, ophthalmology, urology, rheumatology and dentistry.

5. Holography:

- Holography is a true three-dimensional image recording on film by lasers. Holograms are used for amusement, decoration on novelty items and magazine covers.

26.4.3 Safe Handling of Lasers:

Lasers come in various shapes and forms. They have many uses in teaching, research, manufacturing, medicine, dentistry, communications, and shop checkouts and most commonly at work in the office. In fact, some applications may be so well engineered that users are not even aware that the equipment contains a laser.

Laser beam hazards:

The health effects that could occur due to exposure to a laser beam are damage to the skin and eyes. Skin effects include erythema, elastosis (photoageing), and immediate pigment darkening (tanning), burns and skin cancer. Eye effects include photokeratitis, photoconjunctivitis, cataracts, photoretinal damage and burns. The nature of safety measures which are generally taken to avoid the hazards or damages can caused by the lasers are adopted and procedure according to the use and intensity of the laser. Following are some general safety protocols which have to be taken into consideration while working with lasers.

- Only **trained, authorized personnel** may operate lasers. Authorization is received from the authorized laser user and the Laser Safety Officer.
- **NEVER** put yourself into any position where your eyes approach the axis of a laser beam (even with eye protection on).
- **Keep beam paths below or above** standing or sitting eye level. Do not direct them towards other people.
- **Do not damage laser protective housings**, or malfunctioned the interlocks on these housings.
- **Eliminate all reflective material** from the vicinity of the beam paths.
- **Never use viewing instruments** to look directly into a laser beam. If this is necessary, install an appropriate filter into the optical element assembly.
- **Keep ambient light levels as high** as operations will permit.
- **Do not work alone** when performing high power laser operations.

Self-Assessment Questions:

1. How bright is a laser beam when viewed from the side?
2. How do you focus regular light to make it a laser beam?
3. Why do we call laser as a non-material knife?

SUMMARY

- Atomic spectra deal with the spectrum of frequencies of electromagnetic radiation emitted or absorbed by an electron during transitions between different levels of energy within an atom.
- An optical spectrum is obtained, when light passes through a prism or a diffraction-grating.
- A line of particular frequency or wavelength emitted or absorbed by atoms is called spectral line.
- The spectra for each element are unique because each element contains differing numbers of electrons and thus different energy levels.
- Niels Bohr explained the atom of hydrogen. Hydrogen atom consist of proton and electron revolving around the nucleus
- According to Bohr's the radii of Hydrogen atom have the quantized values $r = \frac{n^2\varepsilon_0 h^2}{\pi m e^2}$ or $r = r_0 n^2$ where $r_0 = 5.29 \times 10^{-11} m = 0.53 A^0$ is called Bohr's first orbit.
- X-Rays are powerful electromagnetic rays of higher energy and can pass through most objects, including the body.
- X-rays are produced when fast electrons, or cathode rays, strike a target, such as the walls or anode of a low-pressure discharge tube.
- The light from a laser is quite intense, monochromatic, and one directional.
- The process of photon emission by an excited atom with no external influences is called spontaneous emission.
- Radiation which occurs as a result of external exposure is called induced or stimulated emission.
- Metastable state is a particular excited state of an atom, nucleus, or other system that has a longer lifetime than the ordinary excited states.
- The process of making the population of atoms in the higher energy state (Excited state) more than that of lower energy state (ground state) is known as population inversion.
- A process in which light is used to raise (pump) electrons from a lower energy level in an atom or molecule to a higher energy level is called optical pumping.
- The helium-neon laser was the first continuous wave (CW) laser ever constructed.
- Laser light is Monochromatic, Highly Intense, Coherence and single directional.
- Lasers are widely used in tools, communication, defence, medicine and holography.
- The health effects that could occur due to exposure to a laser beam are damage to the skin and eyes.
- The nature of safety measures which are generally taken to avoid the hazards or damages can caused by the lasers are adopted and procedure according to the use and intensity of the laser.

EXERCISE

Section (A): Multiple Choice Questions (MCQs)
Choose the correct answer:

1. Atomic spectra are also known as:
   (a) Discrete spectra (b) Line spectra
   (c) Emission spectrum (d) Continuous spectra
2. The ratio of kinetic energy to total energy for an electron in a Bohr orbit is:
   (a) 1:1 (b) 2:3 (c) 1:2 (d) 2:1
3. The Bohr radius increases as the principal quantum number:
   (a) increases (b) decreases
   (c) remains constant (d) oscillates
4. Laser beams consist of:
   (a) Highly coherent electrons (b) Highly coherent photons
   (c) Highly coherent phonons (d) Highly coherent neutrons
5. The ruby laser is an example of:
   (a) Optical pumping (b) Electrical pumping
   (c) Chemical pumping (d) Thermal pumping
6. Laser action requires a medium with at least:
   (a) Three energy levels (b) Four energy levels
   (c) Two energy levels (d) Five energy levels
7. Population inversion occurs in:
   (a) Active medium (b) Passive medium
   (c) Gaseous medium (d) Vapour medium
8. X-rays transfer _____ to metals.
   (a) Energy (b) Force (c) Pressure (d) Momentum
9. X-rays are deflected by:
   (a) Magnetic fields (b) Electric fields
   (c) Gravitational fields (d) No fields
10. Doubling the voltage of an X-ray tube:
   (a) Halves the intensity (b) Keeps the intensity unchanged
   (c) Doubles the intensity (d) Quadruples the intensity

Section (B): CRQs (Short Answered Questions):

1. Why do different elements have different spectra?
2. In the Bohr model, how many times larger is the radius of the fifth Bohr orbit compared to that of the first Bohr orbit?
3. What is the difference between X-Rays and Gamma Rays?
4. Sate the properties of X-Rays, which makes it possible to detect cracks in bones.

5. What is the energy of a photon that, when absorbed by a hydrogen atom, could cause an electronic transition from (a) the n =3 state to the n = 5 state and (b) the n = 5 state to the n = 8 state?
6. What are the (a) wavelength range and (b) frequency range of the Lyman series and the Balmer series?
7. Distinguish between spontaneous and stimulated emission.
8. Explain why population inversion is necessary in a laser?
9. In an optically pumped laser, the light that causes optical pumping is always shorter in wavelength than the laser beam. Explain
10. A hydrogen atom is in its first excited state (n = 2). Calculate (a) the kinetic energy of the electron, (b) the potential energy of the system, and (c) the total energy of the system.

Section (C): ERQs (Long Answered Questions):

1. What are the postulates of Bohr's Model of hydrogen atom? Discuss the importance of this model to explain various series of line spectra in hydrogen atom. Do any of the assumptions of the Bohr's theory of hydrogen atom contradict with the classical Physics? Derive the expression for total energy of electron in nth Bohr orbit and show that- $E_n \propto 1/n^2$.
2. How X-rays are produced? State the purpose of cooling fins in the X-ray tube.
3. Explain why X-rays are appropriate in study of crystalline structure material? Write some main properties of X-rays.
4. What is Laser? Write the characteristics of Laser light. Can a two-level system be used for the production of Laser? Why?
5. What is pumping? What are the different methods of pumping? Explain optical pumping.
6. What is the principle of Laser? Write the construction and working of Helium-neon laser.
7. Give some important properties of lasers. Also write the uses of lasers in the field of medicine, defense and communication.
8. What is wave number? Derive the expression $\frac{1}{\lambda} = R_H\left(\frac{1}{p^2} - \frac{1}{n^2}\right)$.
9. Derive the expression for Bohr's radius and develop a general relation for radii of quantized orbits of hydrogen atom.

Section (D): **Numerical:**

1. Calculate the energy of an electron in the n = 2 orbit of a hydrogen atom according to the Bohr model. **(Ans. $-5.447 \times 10^{-19}\ J$)**
2. Calculate the speed of the electron if it orbits in(a) the smallest allowed orbit and (b) the second smallest orbit? (c)If the electron moves to larger orbits, does its speed increase, decrease, or stay the same? **(Ans. (a) $2.19 \times 10^{6} m/s$ (b) $1.09 \times 10^{6} m/s$ (c) the speed of the electron will decrease)**
3. What are the (a) energy, (b) magnitude of the momentum, and (c) wavelength of the photon emitted when a hydrogen atom undergoes a transition from a state with n = 3 to a state with n = 1? **(Ans. (a) 12.1 eV (b) 6.45 X 10^{-27} N.s (c) 102 nm)**
4. What is the energy of the photon emitted by hydrogen atom when the hydrogen atom changes directly from the n = 5 state to the n = 2 state? **(Ans. 2.85 eV)**
5. How much work must be done to pull apart the electron and the proton that make up the hydrogen atom if the atom is initially in (a) its ground state and (b) the state with n =3? **(Ans. (a) 13.6 eV (b) 1.51 eV)**
6. (a) What is the wavelength of light for the least energetic photon emitted in the Balmer series of the hydrogen atom spectrum lines?
   (b) What is the wavelength of the series limit? **((a) 659 nm (b) 366 nm)**
7. A laser emits light with a wavelength of 632.8 nm and has a power output of 55 mW. Calculate the energy of one photon emitted by this laser. **(Ans. 3.14 X 10^{-19} J)**
8. Calculate the wavelength of X-rays if the energy of one photon emitted by the X-ray machine is 1.9878×10−15 joules. **(Ans. 0.1 nm)**

Introduction:

In this chapter, we are going to explore nuclear physics. This field explores the properties and behavior of atomic nuclei, including radioactivity, nuclear reactions, and nuclear energy. First, we will learn about isotopes, which are like different versions of the same atom also see how scientists use mass spectrometers to find these isotopes. Then, we will learn radioactivity, understanding why some atoms emit radiation. Finally, explore the mysteries of nuclear reactions and the cool processes that make nuclei change, even discover how stars and the sun make energy, explore how we can create affordable nuclear energy with reactors, and see how nuclear radiation is used in our everyday lives.

DO YOU KNOW?

Henri Becquerel discovered uranium's invisible energy, sparking curiosity. Rutherford's bold experiment revealed the dense nucleus at the atom's center, like a strong bowling pin, holding most of the atom's weight and binding it together with the powerful nuclear force

27.1 Atomic Nucleus and Isotope:

The atom is the building block of matter; an atom is composed of a positively charged nucleus, with a cloud of negatively charged electrons surrounding it, bound together by electrostatic force. and atomic nucleus is the small, dense region consisting of protons and neutrons at the center of an atom. Almost all of the mass of an atom is located in the nucleus, with a very small contribution from the electron cloud.

All properties of a nucleus are determined by the number of protons and neutrons it has. A specific combination of protons and neutrons is called a nuclide and is a unique nucleus. The following notation is used to represent a particular nuclide.

$$^{A}_{Z}X \ldots\ldots\ldots\ldots 27.1$$

Where X shows any nuclide, Z is atomic number and A is mass number. Table 1 lists these quantities and the symbols commonly used to represent them.

For example, a nuclide of aluminum has a mass number (A) of 27 and an atomic number (Z) of 13. Therefore, it has 13 protons and 14 neutrons (27 − 13 = 14) as shown in figure.27.1 and equation

$$A = Z + N \ldots\ldots\ldots\ldots\ldots 27.2$$

$$A = Z + N$$

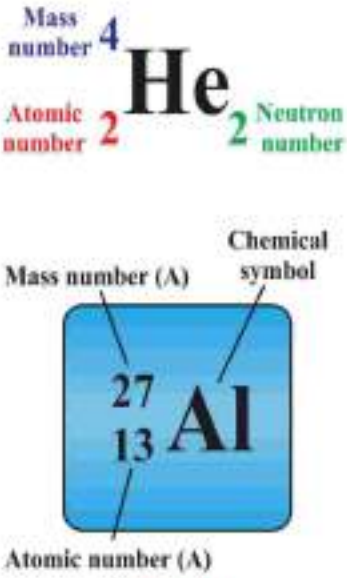


Figure 27.1
Element and its nuclides numbers

Isotopes:

Two or more forms of the same element that contain equal numbers of protons but different numbers of neutrons in their nuclei, and hence they differ in relative atomic mass but not in chemical properties.

Isotopes differ in relative atomic mass but not in chemical properties. For example, carbon has an atomic number of 6. This means that all carbon atoms have 6 protons.
However, carbon has three naturally occurring isotopes: carbon-12, carbon-13, and carbon-14. Carbon-12 has 6 neutrons, carbon-13 has 7 neutrons, and carbon-14 has 8 neutrons.

Isotopes are of two types:

- ***Stable isotopes:*** They do not show any radioactivity and are stable. For example, $_7N^{14}$ and $_7N^{15}$.
- ***Radioactive isotopes:*** They are unstable and show radioactivity. Some of the radioactive isotopes are $_6C^{14}$ for carbon, and $_{19}K^{40}$ for potassium.

Isotopic variations: different numbers of neutrons:

An element can exist in various isotopic forms, each with a different number of neutrons. Elements with low atomic numbers tend to have fewer isotopes, while elements with high atomic numbers tend to have more isotopes.
Every element on the periodic table naturally occurs in various isotopic forms, although some may be more prevalent than others. For instance:

- **Oxygen:** About 99.76% of oxygen in nature is Oxygen-16 (with 8 neutrons). However, Oxygen-17 and Oxygen-18 are also found in much smaller quantities.
- **Uranium:** The naturally occurring isotopes of uranium are Uranium-238 (99.3%) and Uranium-235 (0.7%). Uranium has 35 known isotopes. The most stable isotope of uranium is uranium-238, which has 146 neutrons.
- **Hydrogen:** Hydrogen has only three isotopes: protium, deuterium, and tritium. Protium has no neutrons, deuterium has one neutron, and tritium has two neutrons.

DO YOU KNOW?

The Isotopes have same Physical Properties, Despite having different masses like the isotopes of hydrogen have the same physical properties, such as:

- ✓ Boiling point
- ✓ Melting point:
- ✓ Density
- ✓ Solubility:
- ✓ Color: Colorless
- ✓ Odor: Odorless
- ✓ Taste: Tasteless

Self-Assessment Questions:

1. How many protons are there in the nucleus $^{197}_{79}Au$?
2. How do isotopes of an element differ from each other?
3. Give an example of two isotopes of the same element.

27.1.4 Mass Spectrograph:

A mass spectrograph is a device that can be used to separate isotopes of an element based on their mass and it works by accelerating charged particles through a magnetic field. The particles are then deflected by the magnetic field, and the amount of deflection is determined by the mass of the particle.

Bainbridge mass spectrograph is a type of mass spectrometer that uses a combination of electric and magnetic fields to separate Ions according to their charge-to-mass ratio. It is named after its inventor, Kenneth T. Bainbridge, who developed it in 1933.

The nucleus has a constant probability, ie. the same chance, of decaying in a given time. Therefore, with large numbers of nuclei, it is possible to statistically predict the behavior of the entire group.

27.2.4 Activity (A):

Sometimes we are more interested in the decay rate A ($A = -\frac{\Delta N}{\Delta t}$) than in N itself. The **activity A** of a radioactive sample is the number of disintegrations (decays) occurring per unit of time.

If a radioactive sample contains N atoms at any time t, then its activity at time t is given as

$$A = -\frac{\Delta N}{\Delta t} \ldots\ldots\ldots 27.19$$

where negative sign shows that activity decreases with time. According to law of radioactive decay

$$\frac{\Delta N}{\Delta t} = -\lambda N$$

Therefore,

$$A = \lambda N$$

Since $N = N_0 e^{-\lambda t}$, so

$$A = \lambda N_0 e^{-\lambda t}$$

Putting $A_o = \lambda N_o$ in above equation, we get

$$A = A_0 e^{-\lambda t} \ldots\ldots\ldots 27.20$$

The equation $A = \lambda N$ *and* $A = A_0 e^{-\lambda t}$ are alternative forms of the law of radioactive decay. The total decay rate A of a sample of one or more radionuclide is called the activity of that sample. The SI unit for activity is the Becquerel (Bq), which corresponds to one disintegration per second. Another common unit is the curie (Ci), where 1 Ci is approximately equal to 3.7×10^{10} Bq.

Worked Example 27.2

A radioactive sample of radium ($^{226}_{88}Ra$) contains 3×10^{16} nuclei. The decay constant λ is $1.4 \times 10^{-11} s^{-1}$. What is the activity A of the sample? Convert answer in curies.

Solution:

Step 1: Write down the known quantities and quantities to be found.

Number of nuclei $N = 3 \times 10^{16}$, $\lambda = 1.4 \times 10^{-11} s^{-1}$

Unknown: A = ?

The activity A of the sample is given by:

$$A = \lambda N$$

Step 2: Putting the values,

$$A = 1.4 \times 10^{-11} \times 3 \times 10^{16} = (1.4 \times 3) \times 10^{-11+16} = 4.5 \times 10^5$$

$$A = 4.5 \times 10^5 \text{Bq}$$

Step 3: In order to convert above value in curies, we will divide it by the factor 3.7×10^{10} *Bq/Ci:*

$$A = \frac{4.5 \times 10^5 Bq}{3.7 \times \frac{10^{10} Bq}{Ci}} = 1.1 \times 10^{-5} Ci$$

Result: A=$1.1 \times 10^{-5} Ci$

27.2.6 Exponential Nature of Radioactivity:

Radioactive decay follows an exponential decay pattern. This means that the rate of decay proportionally decreases with the remaining amount of radioactive material and it can be visualized graphically by decay curve.

A decay curve is a plot of the number of radioactive parent nuclei remaining in a sample as a function of half-life. A typical decay curve for a radioactive sample is shown in Figure 27.3. After each half-life, half the remaining parent nuclei have decayed. This is represented in the circles to the right of the decay curve. The blue spheres are the parent nuclei (carbon-14), and the red spheres are daughter nuclei (nitrogen-14). Notice that the total number of nuclei remains constant, while the number of carbon atoms continually decreases over time.

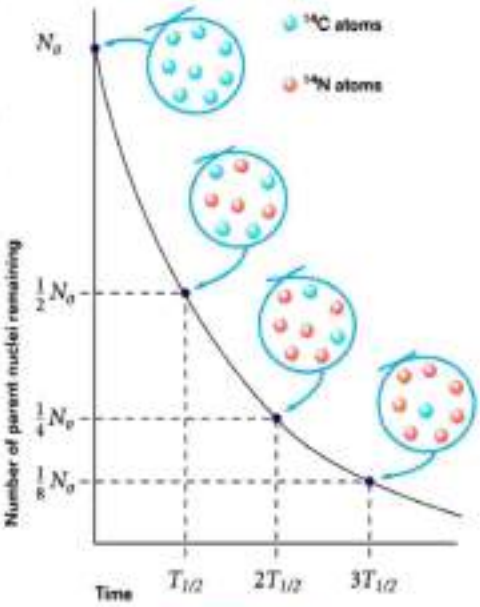


Figure: 27.3
decay curve for a radioactive sample

For example, the initial sample contains 8 carbon-14 atoms. After one half-life, there are 4 carbon-14 atoms and 4 nitrogen-14 atoms. By the next half-life, the number of carbon-14 atoms is reduced to 2, and the process continues. As the number of carbon-14 atoms decreases, the number of nitrogen-14 atoms increases.

The curve gets shallower and shallower as time goes on. This is because there are fewer and fewer radioactive nuclei left to decay, so the rate of decay slows down. This exponential decay is a characteristic feature of radioactive substances and is commonly used in radiocarbon dating and other applications in nuclear physics.

The key features of the exponential nature of radioactive decay include:

- The rate at which radioactive atoms decay is proportional to the quantity of the substance present.
- When plotted on a graph, the decay curve is a smooth, continuous curve that approaches but never reaches zero.
- The curve demonstrates a steeper decline at the beginning (short times) and a gradual decrease as time progresses.
- While we can't predict when a specific atom will decay, we can predict the probability of decay within a certain time frame.

27.2.7 Half-Life (t1/2):

The half-life of a radioactive substance is the time it takes for half of the initial amount of the substance to decay or transform into another element. It's a measure of the stability of an atom and the rate at which it loses its radioactivity.

Decay Constant (λ):

The decay constant, is a measure of the rate at which a radioactive substance decays. It's denoted by the symbol λ and represents the probability of decay per unit time.

Formula:

The half-life formula is:

$t1/2 = \ln(2) / \lambda$

Where:

- t1/2 is the half-life
- ln(2) is the natural logarithm of 2 (approximately 0.693)
- λ is the half-life constant (decay constant)

This shows that the half-life is inversely proportional to the decay constant. A higher decay constant means a shorter half-life, and vice versa.

Worked Example 27.3

A 50.0-g sample of carbon is taken from the pelvis bone of a skeleton and is found to have a carbon-14 decay rate of 200.0decays/min. It is known that carbon from a living organism has a decay rate of 15.0 decays/min-g and that C-14 has a half-life of 5730 years = 3.01×10^{9} min. Find the age of the skeleton.

Solution:

Step 1: Write down the known quantities and quantities to be found.

Given:A = 200 decays/min

Calculate the original activity A_0 from the decay rate and the mass of the sample:

$A_0 = 15$ decay/min-g $\times$ 50g $= 7.5 \times 10^2$ decays/min

$T_{1/2} = 3.01 \times 10^9$min

Unknown: t = ?

Step 2: Age of skeleton can be found by the equation:

$$A = A_0 e^{-\lambda t}$$

Re-arranging,

$$t = \frac{\left(\ln \frac{A}{A_0}\right)}{\lambda}$$

Step 3: Find the decay constant from the half-life:

$$\lambda = \frac{0.693}{T_{1/2}} = \frac{0.693}{3.01 \times 10^9} = 2.3 \times 10^{-10} min$$

27.3.1 Unified Mass Scale and Mass Defect:

Unified mass scale:

It is also known as the **Unified Mass Scale** or **Atomic Mass Unit (amu)**, is a unit of measurement used in atomic and nuclear physics to express the mass of atoms and molecules on a scale that is convenient for working with atomic and molecular masses based on the carbon-12 isotope.

The Atomic Mass Unit (a.m.u) is defined as 1/12th of the mass of $_6C^{12}$ atom (C−12). According to Avogadro's Hypothesis, the number of atoms in 12 gm of $_6C^{12}$ is equal to Avogadro number i.e., 6.023×10^{23}.

The mass of one carbon atom $= \frac{12}{6.023\times10^{23}} = 1.992678 \times 10^{-26}$kg.

Therefore 1 a.m.u is defined as:

$1\text{a.m.u} = \frac{1}{12} \times 1.992678 \times 10^{-26} = 1.660565 \times 10^{-27}\text{kg}$

For examples:

- Mass of a proton $m_p = 1.007276$ a.m.u$= 1.67865 \times 10^{-27}$kg
- Mass of a neutron$m_n = 1.008665$ a.m.u$= 1.67495 \times 10^{-27}$kg
- Hydrogen atom $_1H^1 = 1.00784$ a.m.u
- Chlorine atom = 35.47 a.m.u

Energy equivalent of a.m.u:

According to Einstein mass-energy relation, the energy equivalent to mass m given by: $E = mc^2$ where c is velocity of light.

Let mass m = 1 a.m.u = 1.680665×10^{-27}kg, $c = 3 \times 10^8$ m/s. Then energy equivalent to 1 a.m.u is given as:

$1\text{ a.m.u} = 1.680666 \times 10^{-27} \times (8 \times 10^8)^2 = 1.4925 \times 10^{-10}$ Joules

$1\text{ a.m.u} = \frac{1.4925\times10^{-10}}{1.602\times10^{-13}} = 931.5$ MeV

Where 1MeV $=1.602\times10^{-13}$J. Energy equivalent to mass of electron, proton and neutron are:

$m_e = 0.511$MeV, $m_p = 938.279$MeV and $m_n = 939.573$MeV.

The mass and rest energy of all particles are given in the table 27.2.

Table 27.2 Mass and Rest energy of atomic particles

Particle	m (kg)	m (u)	E_R (MeV)
Proton	1.673×10^{-27}	1.007276	938.3
Neutron	1.675×10^{-27}	1.008665	939.6
Electron	9.109×10^{-31}	1.000549	0.5110

Binding Energy and Mass Defect:

Binding energy is the energy required to separate all the protons and neutrons in a nucleus as shown in figure 27.5. It is a measure of the stability of the nucleus. The mass defect, which is the difference between the mass of the nucleus and the sum of the individual masses of its protons and neutrons, is converted into binding energy according to Einstein's equation $E = mc^2$.

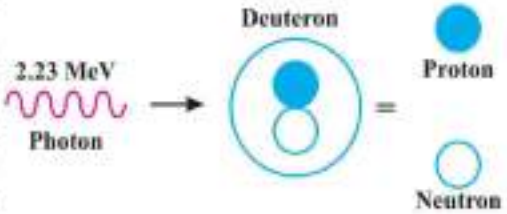


Figure 27.5 Splitting of nucleus

This energy is what holds the nucleus together, overcoming the repulsive forces between the positively charged protons. The mass of the nucleus so formed is less than the sum of the masses of the constituent protons and neutrons. This mass difference is called **mass defect** and is denoted by Δm.

If Z is the number of protons in the nucleus (Atomic number Z) and A is the atomic mass, then the number of neutrons in the nucleus is (A - Z). If m_p is the mass of the proton and m_n is the mass of the neutron, then:

Sum of the masses of the protons and neutrons = $Zm_p + (A - Z)m_n$.

If M_N is the actual mass of the nucleus, then Mass defect Δm is given by,

$$\Delta m = Zm_p + (A - Z)m_n - M_N \quad \ldots\ldots\ldots 27.21$$

If mass defect is Δm then the binding energy of the nucleus according to Einstein mass energy relation is given as:

Binding Energy = $E_B = \Delta mc^2$

$E_B = c^2[Zm_p + (A-Z)m_n - M_N]$

where M_N is the mass of the nucleus, m_p the mass of the proton and m_n the mass of the neutron.

Worked Example 27.5

Calculate the mass defect and binding energy of deutron which contains one proton and one neutron. The mass of deutron nucleus is 2.0136 a.m.u, mass of a proton is 1.0073 a.m.u and mass of a neutron is 1.0087 a.m.u.

Solution:

Step 1: Write down the known quantities and quantities to be found.

mass of proton: 1.0073 a.m.u, mass of neutron = 1.0087 a.m.u and mass of deutron = 2.0136 a.m.u.

Unknown: $\Delta m = ?$, $E_B = ?$

Step 2: The combined mass is of proton and neutron = [1.0073 + 1.0087] = 2.1060 a.m.u.

Step 3: Therefore, mass defect: $\Delta m = 2.1060 - 2.0136 = 0.0024$ a.m.u.

The biding energy is given by:

$$E_B = \Delta m \times 931.5 = 0.0024 = 2.2356 MeV$$

Result: $E_B = 2.2356 MeV$

27.3.2 Binding Energy Per Nucleon:

The binding energy per nucleon ($\overline{E_B}$) is the average energy required to release a nucleon from the nucleus. It is given mathematically as

$$\overline{E_B} = \frac{E_B}{A} \ldots\ldots\ldots 27.22$$

where A is the total number of nucleons in the nucleus.

The binding energy per nucleus$\overline{E_B}$ of a nucleus is more important than total binding energy E_B of the nucleus because $\overline{E_B}$determines the stability of the nucleus. Higher is the value of the binding energy per nucleon, more stable is the nucleus. The nuclei with less binding energy per nucleon are comparatives less stable.

The graph between binding energy per nucleon and mass number of different nuclei is shown in Figure 27.6. Following are the main features of binding energy curve:

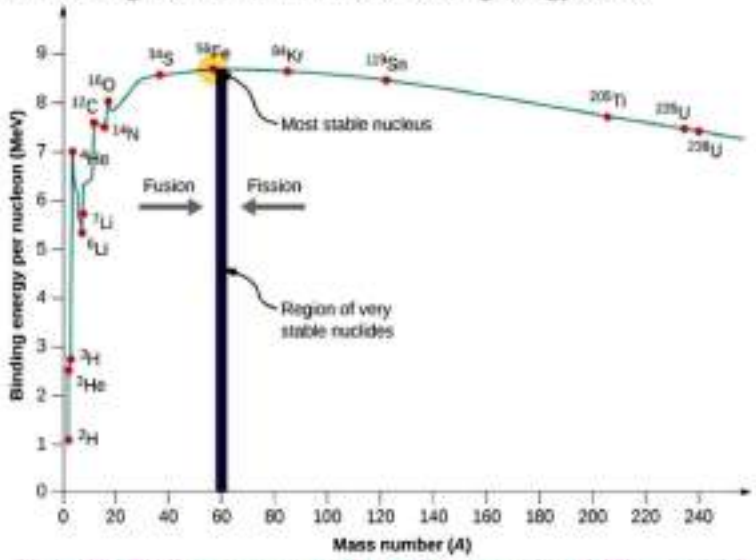


Figure 27.6 Binding energy per nucleon and mass number of different nuclei

- The binding energy per nucleon for light nuclei (i.e. $_aH^2$) is very small. The binding energy per nucleon increases rapidly for nuclei up to mass number 20.
- The curve possesses peaks corresponding to nuclei $_2He^4$,$_4Be^8$, $_6C^{12}$, $_8O^{16}$ and $_{10}Ne^{20}$.
- The peaks indicate that these nuclei are more stable than other nuclei in their neighborhood. ù After mass number 20 binding energy per nucleon increases gradually.
- The curve has average value of binding energy per nucleon of about 8.6 MeV for a very considerable range of mass number 10 to 120. In this range the curve is more or less flat.
- For mass number A = 6(i.e., $_{26}F^{56}$), the binding energy per nucleon is maximum and is equal to 8.8 MeV.

1. **External radiation** exposure occurs when a person is exposed to radiation from a source outside the body, such as an X-ray machine or a nuclear power plant.
2. **Internal radiation** exposure occurs when a person ingests or inhales radioactive material, which can then accumulate in the body and emit radiation.

Radiation exposure can also be measured in other units, including sieverts (Sv), millisieverts (mSv), and microsieverts (μSv).

Biological Effects of Radiation:

When radiation is absorbed by matter, especially living tissue, it can induce significant changes. The biological effects of radiation are diverse and depend on several factors, including:

- **Type of radiation:** Each type interacts differently. Alpha particles, while unable to penetrate deeply, cause disorder within cells they reach. Beta particles travel farther but deposit less energy, while gamma rays can pierce through tissues, potentially affecting multiple organs.
- **Dose absorbed:** This quantifies the energy deposited per unit mass of tissue. Higher doses generally translate to more pronounced effects.
- **Duration of exposure:** Acute (short-term) exposure like an X-ray differs from chronic (long-term) exposure from environmental sources or occupational hazards.
- **Individual sensitivity:** Age, health status, and genetic predispositions can influence susceptibility.

The biological effects of radiation include:

- **Acute radiation syndrome (ARS):** High-dose exposure can trigger ARS, a complex condition affecting multiple organ systems with symptoms like nausea, vomiting, hair loss, and bone marrow suppression.
- **Cancer risk:** Chronic low-dose or high-dose exposures can increase the risk of various cancers, depending on the affected tissue.
- **Genetic effects:** Germ cell mutations can lead to congenital disabilities in offspring.
- **Reproductive issues:** Fertility can be impaired, and miscarriage risk may increase.
- **Developmental effects:** Early exposure inutero can affect fetal development, causing physical and cognitive impairments.

27.5.2 Dosimetry:

Dosimetry is the science of measuring and quantifying radiation dose. It is a field of physics that deals with the interaction of radiation with matter and the biological effects of radiation exposure.

Dosimetry is derived from the Greek word 'dos', meaning 'gift' or 'a giving'. In the context of physics, it represents the scientific discipline that quantifies the amount of radiation 'given' to or absorbed by an object or body. The goal is to determine potential biological effects, ensuring safe limits and monitoring radiation exposure.

Key Components of Dosimetry:

1. **Absorbed Dose:** The amount of radiation energy absorbed per unit mass of the material. Measured in Gray (Gy), it is a pivotal metric in understanding radiation's effect on tissues.

2. **Equivalent Dose:** Not all radiation types have the same biological impact, even if their absorbed doses are identical. To account for this, radiation weighting factors are introduced to derive the equivalent dose. Measured in Sievert (Sv), it provides a more biologically relevant dose metric.
3. **Effective Dose:** Considering that different tissues have varying sensitivities to radiation, tissue weighting factors are used. The effective dose, also in Sieverts (Sv), offers a measure that summarizes the potential overall harm to the whole organism.
4. **Operational Dose Quantities:** These are practical quantities used for routine monitoring in radiological protection. They are designed to estimate effective dose or equivalent dose to a particular tissue. Examples include ambient dose equivalent and personal dose equivalent.

Dosimetry is performed using devices known as **dosimeters**, which can be worn by people working with or around radioactive materials to monitor their exposure levels. These devices can measure and record the dose of radiation over time.

27.5.3 Medical Uses of Nuclear Radiation:

Medical use of nuclear radiation is quite common in today's hospitals and clinics. It contains various diagnostic and therapeutic applications that control the properties of ionizing radiation to diagnose and treat diseases. Some examples include:

Imaging:

- **X-rays:** The workhorse of medical imaging, X-rays utilize electromagnetic radiation to reveal fractures, bone structures, and internal injuries.
- **CT scans:** Combining multiple X-ray images, CT scans provide detailed cross-sectional views of organs and tissues, aiding in diagnosing tumors, infect lons, and internal bleeding.
- **PET scans:** Positron emission tomography (PET) utilizes radioactive tracers to map metabolic activity within the body, uncovering cancerous tumors and other metabolic disorders.

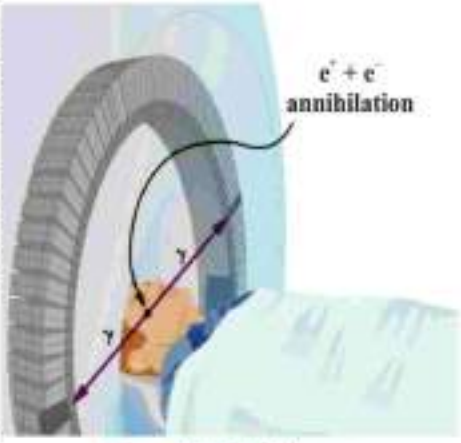


Figure 27.13 The CT scan

Nuclear medicine procedures:

- **Bone scans:**
  Radioactive tracers identify bone diseases like osteoporosis and cancer metastases.
- **Thyroid scans:**
  Radioactive iodine helps diagnose and treat thyroid disorders.
- **Lung scans:**
  Technetium-99m helps assess lung function and detect blood clots.

Treatment:

- **Radiotherapy:** Harnessing the destructive power of radiation, targeted beams are used to shrink and destroy cancerous tumors with remarkable precision. This non-invasive therapy plays a crucial role in treating various cancers, often in conjunction with surgery and chemotherapy.
- **Brachy therapy:** Radioactive implants placed directly within tumors deliver high doses of radiation locally, minimizing damage to surrounding tissue. This approach is particularly effective for treating prostate, cervical, and head and neck cancers.
- **Radioisotope therapy:** Radioactive isotopes are used to treat specific conditions like hyperthyroidism (radioactive iodine) and blood disorders (radioactive phosphorus).

Other Applications:

- **Pain management:** Radiofrequency ablation utilizes radio waves to heat and destroy nerve tissue, offering pain relief for chronic conditions like back pain and arthritis.
- **Sterilization:** Medical instruments and equipment are sterilized using gamma radiation, ensuring sterility and preventing infections.

27.5.4 Exposure of radiation:

This process creates charged particles (Ions) and free radicals, which can disrupt molecular structures and biological processes. Examples of ionizing radiation include:

- **X-rays:** Electromagnetic radiation produced by high-energy electron beams or X-ray tubes. X-rays are commonly used in medical imaging, security screening, and industrial applications.
- **Gamma rays:** High-energy electromagnetic radiation emitted by radioactive decay processes, such as those occurring in radioactive isotopes like cobalt-60 and cesium-137. Gamma rays are used in medical imaging (gamma cameras), radiation therapy, and sterilization processes.
- **Alpha particles:** Alpha particles are emitted during the decay of heavy elements like uranium and radium and have low penetrating power but can cause significant damage if inhaled or ingested.
- **Beta particles:** Beta particles can penetrate deeper into tissues than alpha particles but are less damaging. They are used in medical imaging (positron emission tomography) and radiation therapy.
- **Neutrons:** Neutrons can induce nuclear reactions and are used in neutron activation analysis, neutron radiography, and certain types of cancer therapy.
- **Cosmic rays:** High-energy particles from outer space continuously bombard Earth's atmosphere, generating secondary radiation that reaches the surface.
- **Industrial applications:** Gauges, sterilization processes, and smoke detectors often employ ionizing radiation for various industrial purposes.

These forms of ionizing radiation have various applications in medicine, industry, research, and other fields but require careful handling and monitoring to minimize risks to human health and the environment.

27.5.5 Uses of Radioactive Tracers:

One of the most important uses of nuclear radiation is the location and study of diseased tissue. This can be done by radioactive tracers. **Radioactive tracers** are radioactive substances that are used to track the movement and behavior of specific molecules or compounds in various systems. These tracers emit radiation that can be detected through biochemical reactions, metabolic pathways, fluid flow, and environmental transport. Some examples of the uses of radioactive tracers include:

Medical Imaging:

- **Positron Emission Tomography (PET):** Radiotracers labeled with positron-emitting isotopes (e.g., fluorine-18, carbon-11) are injected into the body and used to visualize metabolic activity, blood flow, and receptor binding in tissues. PET imaging is valuable for diagnosing and monitoring diseases such as cancer, heart disease, and neurological disorders.

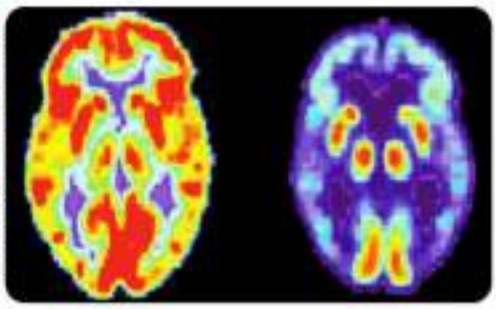


Figure 27.14 Positron Emission Tomography

Industrial Processes:

- **Flow Visualization:** Radiotracers are injected into fluids or gases to track flow patterns and detect leaks or blockages in industrial pipelines, heat exchangers, and reactors. This technique is used in industries such as petrochemicals, food processing, and nuclear power.
- **Process Optimization:** Radiotracers are used to monitor and optimize chemical reactions, mixing processes, and material transport in industrial processes. By tracking the movement of tracers, engineers can identify bottlenecks, improve efficiency, and ensure product quality.
- **Soil Fertility and Nutrient Uptake:** Radioactive isotopes like phosphorus-32 (^{32}P) are used to study nutrient uptake in plants. By tagging fertilizers with these tracers, scientists can track how nutrients move through the soil and are absorbed by plants. This information helps in optimizing fertilizer application, ensuring that crops receive the right amount of nutrients.
- **Environmental Impact Assessments:** Radioactive tracers help in studying the environmental impact of agricultural practices. For instance, they can be used to trace the movement of pollutants or contaminants from agricultural fields to water bodies, enabling better management practices to protect the environment.

EXERCISE

Section (A): Multiple Choice Questions (MCQs)

Choose the correct answer:

1. Identify particle x in the following nuclear reaction $^{9}_{4}Be + ^{4}_{2}He \rightarrow ^{12}_{6}C + x$
   (a) Electron (b) proton (c) neutron (d) photon
2. In the equation $^{27}_{13}Al + ^{4}_{2}He \rightarrow ^{30}_{15}P + X$,The correct symbol for X is:
   (a) $^{-1}_{0}e$ (b) $^{1}_{1}H$ (c) $^{4}_{2}He$ (d) $^{1}_{0}n$
3. Beta rays emitted by a radioactive material are:
   (a) Electromagnetic radiations (b) The electrons orbiting around the nucleus
   (c) Charged particles emitted by the nucleus (d) Neutral particles
4. The number of α and β particles emitted in the following radioactive decay is:
   $$^{200}_{90}X \rightarrow ^{168}_{80}Y + ^{7}_{?}X$$
   (a) 8 and 6 (b) 6 and 8 (c) 8 and 8 (d) 6 and 6
5. If radium has half-life of 5 years. Thus for a nucleus in a sample of radium, the probability of decay in ten years is:
   (a) 50% (b) 75% (c) 100% (d) 60%
6. Emission of β particles by an element affects its mass number **A** in the following way:
   (a) Increases by 1 (b) decreases by 1
   (c) Increases by 2 (d) remains the same
7. As the number of nucleons in a nucleus increases, the binding energy per nucleon:
   (a) Increases continuously with mass number
   (b) Decreases continuously with mass number
   (c) remains constant with mass number
   (d) First increases and then decreases with increase of mass number
8. Moderator in a nuclear reactor slows down the neutrons to:
   (a) Decrease the probability of escape (b) Increase the probability of nuclear fission
   (c) Decrease the probability of absorption (d) all of the above
9. Emission of β^{+} particles by an element affects its atomic number Z in the following way:
   (a) Increases by 1 (b) decreases by 1
   (c) Increases by 2 (d) remains the same
10. The half life period of a radioactive element is 100 days. After 400 days, 16 g of the element will be reduced to
   (a) 8g (b) 4g (c) 2g (d) 1g

Section (B): CRQs (Short Answered Questions):

1. Why are protons and neutrons necessary for the stability of an atomic nucleus?
2. How do isotopes of an element differ and why are these differences significant?
3. Name the two methods of controlling a chain reaction.
4. Why is nuclear decay described as spontaneous and random?
5. Explain what happens during the nuclear fusion process.

6. How are activity and decay constant related in radioactive materials?
7. What is meant by binding energy and binding energy per nucleon?
8. Why is the concept of half-life important in studying radioactive substances?
9. Explain the process of positron emission.

Section (C): ERQs (Long Answered Questions):

1. What is radioactivity? State the law of radioactive disintegration. Show that radioactive decay is exponential in nature.
2. Explain the main differences between alpha, beta, and gamma emissions.
3. A fission reactor produces energy to drive a generator. Describe briefly how this energy is produced.
4. Define Q-value of a nuclear reaction and its significance.
5. Describe the construction and working of GM counter.
6. Discus nuclear reactions induced by neutrons. Why are neutrons preferred to other particles?
7. What is a nuclear reaction? Explain nuclear fission and fusion.
8. How do the Sun and stars produce energy? What is the proton-proton cycle? Explain with details.

Section (D): Numerical:

1. In 9.0 days the number of radioactive nuclei decreases to one-eighth the number present initially. What is the half-life (in days) of the material? **[Ans: 3 days]**
2. The $^{32}_{15}P$ isotope of phosphorus has a half-life of 14.28 days. What is its decay constant in units of s^{-1}? **[Ans: $5.62\times10^{-7}s^{-1}$]**
3. Find the binding energy (in MeV) for lithium $^{7}_{3}Li$ (atomic mass = 7.016 003 u). **[Ans: 39.2 MeV]**
4. The binding energy of a nucleus is 225.0 MeV. What is the mass defect of the nucleus in atomic mass units? **[Ans: 0.2415 u]**
5. A copper penny has a mass of 3.0 g. Determine the energy (in MeV) that would be required to break all the copper nuclei into their constituent protons and neutrons. Ignore the energy that binds the electrons to the nucleus and the energy that binds one atom to another in the structure of the metal. For simplicity, assume that all the copper nuclei are $^{63}_{29}Cu$ (atomic mass = 62.939 598 u). **[Ans: 1.51249×10^{25}MeV]**
6. Write the β^+ decay process for each of the following nuclei with their proper chemical symbols including Z and A for each daughter nucleus: (a) $^{18}_{9}F$(b) $^{15}_{8}O$.
   [Ans: (a) $^{18}_{9}F \rightarrow {}^{18}_{8}O + e^+ + \nu_e$ (b) $^{15}_{8}O \rightarrow {}^{15}_{7}N + e^+ + \nu_e$]
7. A device used in radiation therapy for cancer contains 0.50 g of cobalt $^{60}_{27}Co$ (59.933 819 u). The half-life of $^{60}_{27}Co$ is 5.27 years. Determine the activity of the radioactive material.
   [Ans: 2.10×10^{13}decays/second (Bq)]

Introduction:

Let's explore how our universe was formed and what holds everything together. People have been curious about these questions for a very long time. Originally, scientists believed that atoms were the smallest pieces that made up everything. But, in the 1800s, they discovered that atoms have smaller parts called electrons, protons, and neutrons.

Today, physics tells us about even tinier pieces that make up everything. These super small pieces are called "elementary" or "fundamental" particles, and they're so simple that they don't have any smaller parts inside them. Molecules, atoms, protons, and neutrons are not fundamental particles because these are made up of other particles, while electrons, quarks, and photons are fundamental particles. Studying all these elementary particles and understanding how they behave is the main goal of particle physics. Imagine these particles as tiny puzzle pieces that collectively form everything around us! To make sense of this cosmic puzzle, scientists have developed various theories. These theories aid in explaining the behavior and interactions of these particles. In this chapter, we will study one of the most crucial theories known as the **Standard Model**.

DO YOU KNOW?

Fundamental particles
Those particles which are simple, structure less, and are not made up of other particles.

Particle Physics is the branch of Modern Physics which deals with the study and search of the ultimate constituents of matter (elementary particles) and their interactions.

28.1 The Standard Model:

Our universe is made up of two things: matter and energy (radiation). To understand them better, scientists have divided particles into two main groups: matter particles and force particles. Scientists have identified many elementary particles belonging to these categories. These particles are categorized and explained in detail in the Standard Model of Particle Physics, which is the best-known theory to date. It is a framework that explains three of the four fundamental forces (electromagnetism, the weak nuclear force, and the strong nuclear force) and all known elementary particles.

The Standard Model classifies all known elementary particles into two main classes:

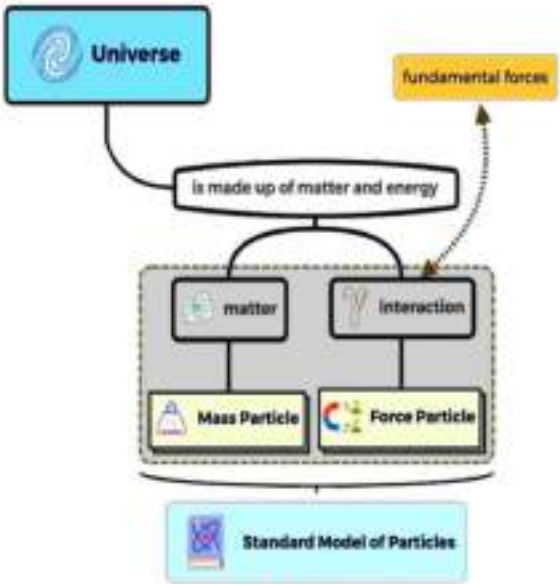


Figure 28.1:
The universe is made up of matter and energy.

Fermions:

These are the matter particles, which make up everything in the universe.

Bosons:

These are the force-carrying particles, which mediate the interactions between fermions.

The Standard Model is a very successful theory. It has been able to predict the existence of many new particles, and it has been confirmed by a wide range of experiments. However, the Standard Model is not perfect. It does not explain gravity, which is the fourth fundamental force. It also does not explain dark matter and dark energy, which make up most of the universe.

28.1.1 Fundamental forces and their field particles:

The building block or brick of all matter is known as an elementary particle, but these particles cannot make up the whole structure of the universe by themselves. Just like a "brick" is the building block of our homes, we need cement as "glue" to give structure to our homes. Similarly, we need "glue" or "force particles" to build our universe. This glue is known as fundamental forces in the Standard Model. Each fundamental force has a field particle associated with it. A field particle is a virtual particle that mediates the fundamental force between two other particles. The Standard Model has kept all field particles in a class of "Boson."

The term "field particles" refers to particles associated with force fields. According to the Standard Model, bosons are often considered field particles because they are linked to force fields. For example, the photon is a field particle associated with the electromagnetic field.

Scientists have grouped all fundamental forces into four basic types. In order of increasing strength, these forces and their associated field particles are described as:

1. Gravitational Force:

It is the force that attracts two masses towards each other. It is the weakest of the four fundamental forces but acts over long distances.

Field Particle: The hypothetical particle associated with gravity is the graviton, although it has not been observed yet. The Standard Model doesn't explain gravity.

2. Weak Nuclear Force:

The weak force is responsible for radioactive decay, where unstable atomic nuclei break down into smaller, more stable nuclei. It's responsible for processes like beta decay and neutrino emission. This force is weaker than electromagnetic and strong nuclear forces but stronger than gravitational force.

Field Particle: There are three field particles associates with weak nuclear force. These are W^+, W^- and Z bosons. These short-lived bosons carry the force over very small distances, explaining the limited range of the weak force.

3. Electromagnetic Force:

This force is responsible for the interactions between charged particles, such as electrons and protons. It includes both electric and magnetic forces. The electromagnetic force

is stronger than both gravitational and weak nuclear forces. It is also a long-range force, similar to the gravitational force.

Field Particle: The field particle of electromagnetic force is photon. It is mass less and chargeless particle. When ever charged particles interact, they exchange photons, causing the attractive or repulsive forces we observe.

4. Strong Nuclear Force:

The strong force binds quarks together to form protons and neutrons, and it holds atomic nuclei together. It is the strongest force among all forces and acts at subatomic levels.

Field Particle: Gluons are the field particles that mediate the strong force between quarks.

28.1.2 Particle zoo:

One way of studying elementary particles is to classify them into different categories based on certain behaviors and then to look for similarities or common characteristics among the classifications. We know that the Standard Model has classified all elementary particles and their interactions into two main groups according to their spins:

- Particles that carry force, called **boson s**having spin in integer values such as 0,1and 2.
- Particles that make up all matter, called **fermions** having spin in odd half integer values such as ½, 3/2 etc.

Bosons:

The arrangement of all elementary particles in the Standard Model constitutes a particle zoo. All bosons can be classified into two types: Elementary Boson and Composite Boson. Composite bosons consist of quark and anti-quark combinations while elementary bosons are not made up of other elementary particles.

Elementary Bosons:

Elementary bosons are the carriers of the fundamental forces in nature. For examples:

- Gluon
- W^+, W^-andZ^0 Bosons
- Photon
- Higgs Boson– *gives mass to all particles.*
- Graviton

The properties of elementary bosons are given in the table.

Table 28.1 The properties of elementary bosons

Force	Boson	Spin	Strength	Mass
Strong	Gluon	1	1	Massless
Electromagnetic	Photon	1	10^{-2}	Massless
Weak	$W_{\pm}$ Z	1	10^{-7}	80.91GeV
Gravity	Gravitation	2	10^{-39}	Massless

Composite Bosons:

Composite bosons are made up of an even number of fermions. For examples, **mesons** are composite bosons that composed of one quark and one antiquark. Mesons are intermediate mass particles that mediate the strong force between nucleons such as protons and neutrons.

Fermions:

All material particles are made up of fermions. We can further divide fermions in to two sub-classes: **elementary fermions** and **composite fermions**.

Elementary Fermions:

Elementary fermions are building blocks of all material particles. There are not made of any other particles. The fermions come in two types: **leptons** and **quarks**.

(a) Leptons:

Leptons are a group of elementary particles that do not experience the strong nuclear force. There are six types or flavors of leptons, which come in three pairs. The pairs are made up of three charged particles named electron, muon, and tau, along with their Partners called neutrinos (charge less). These all six leptons group into three generations:
Generation I, Generation II and Generation III as shown in the table. 28.2 Each generation consists of one pair of leptons (for example, electron with electron neutrino).

Properties of leptons:

- **Electron:** Negatively charged; commonly found in atoms.
- **Muon and Tau:** Heavier counter parts of the electron.
- **Neutrinos:** Electrically neutral; they interact very weakly with matter.
- Leptons interact via weak and electromagnetic forces but not through the strong force.
- Leptons are stable particles and do not undergo decay under normal circumstances.
- Leptons exist alone and donot form groups.

DO YOU KNOW?

- Our universe belongs to Generation I particles. Particles belonging to Generation II and Generation III are heavier than the particles belonging to Generation I.
- The heavier particles of Generations II and III undergo decay processes, transforming into Generation I particles.

Table 28.2 Generations of Leptons and Quarks

	Fermions			
	Charge	**Generation I**	**Generation II**	**Generation III**
Leptons	-1	electron (e)	muon (μ)	tau (τ)
	0	e-neutrino	μ-neutrino	τ-neutrino
Quarks	+2/3	up (u)	charm(c)	top(t)
	-1/3	down (d)	strange (s)	bottom(b)

28.2.1 Wilson cloud chambers:

A Wilson cloud chamber is a type of tracking detector which works on the principle of ionization. It is a class of gas detector that is used in particle physics and nuclear physics to visualize the tracks of subatomic particles such as electrons, positrons, alpha particles, and cosmic rays. It was invented by Scottish physicist Charles Thomson Rees Wilsonin1912 and played a crucial role in the early discoveries of subatomic particles.

Figure 28.9 Wilson cloud Chambers

Working Principle:

The Wilson Cloud Chamber consists of a sealed container filled with a supersaturated vapor, typically water or ethanol. When a particle passes through the chamber, it ionizes the vapor, creating a trail of droplets that condense around the ionized path. This creates a visible cloud-like track that can be photographed and analyzed.

Construction:

The schematic diagram of Wilson cloud chamber is shown in figure 28.10, which consists of a large cylindrical chamber A, with walls and a ceiling made of glass. It contains dust-free air saturated with water vapor. P is a piston working inside the chamber. When the piston moves down rapidly, adiabatic expansion of the air inside the chamber takes place. The piston is connected to a large evacuated vessel F through a valve V. When the valve is opened, the air under the piston rushes into the evacuated vessel F, thereby causing the piston to drop suddenly. The wooden blocks WW reduce the air space inside the piston. Water at the bottom of the apparatus ensures saturation in the chamber. The expansion ratio can be adjusted by altering the height of the piston.

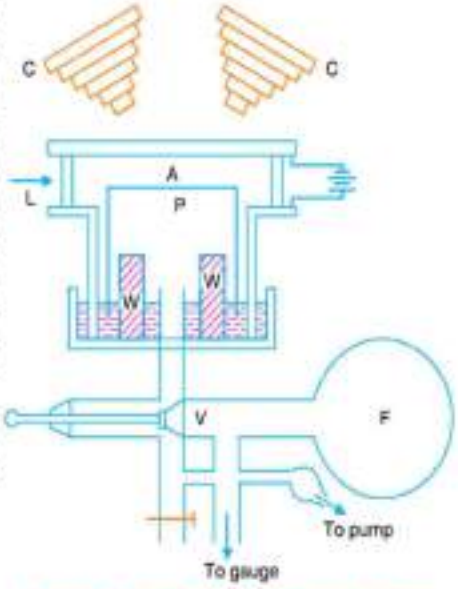


Figure 28.10 Schematic diagram of Wilson Cloud Chamber

As soon as the gas in the expansion chamber is subjected to sudden expansion, the ionizing particles are shot into the chamber through a side window. A large number of extremely fine droplets are formed on all the ions produced by the ionizing particles. These droplets form a track of the moving ionizing particles. At this stage, the expansion chamber is profusely illuminated by a powerful beam of light L and two cameras CC are used to photograph the tracks as shown in figure 28.10. The process of expansion, shooting of the

ionizing particles into the expansion chamber, illuminating the chamber and clicking the camera must all be carried out in rapid succession in order to get satisfactory results. The type of ionizing particle can be identified by its track in the cloud chamber. Alpha particles, being relatively massive, travel in straight, thick, and clearly defined paths. Beta particles, being lighter, are easily deflected and create thin, curved paths. The cloud chamber has been instrumental in discovering many elementary particles, such as the positron and meson.

Use of Wilson Cloud Chambers:

- **Particle Identification:** Wilson cloud chambers were historically instrumental in the identification and study of subatomic particles. By observing the curvature of particle tracks in a magnetic field and the nature of the tracks themselves, scientists could identify and classify various particles.
- **Nuclear Physics Research:** Cloud chambers have been used to study the behavior of particles in nuclear reactions and to investigate the structure of atomic nuclei.
- **Cosmic Ray Studies:** Wilson cloud chambers are also used in cosmic ray research. These instruments can detect and track the passage of cosmic rays, which are high-energy particles originating from space.
- **Education and Outreach:** Cloud chambers are often used as educational tools in physics classrooms and science museums to help students and the general public visualize the behavior of subatomic particles.

Geiger-Muller counter:

A Geiger-Muller counter (GM counter) is a type of gas radiation detector used to measure ionizing radiation. It is a portable and versatile device commonly employed for detecting the presence and intensity of various types of ionizing radiation, including alpha particles, beta particles, and gamma rays. The Geiger-Muller (GM) counter was invented by two German scientists, Hans Geiger and Walther Muller, in 1928.

Figure 28.11 GM Counter

Construction:

The GM counter consists of a hollow metallic chamber as shown in the figure 28.11 that acts as a cathode.

A thin wire anode is also placed along its axis.

The chamber has a sealed window, through which the radiation enters the chamber.

The chamber is filled with an inert gas at low pressure.

There is a counter connected to this system to measure the radiation.

Working:

The chamber is filled with an inert gas (helium, neon, or argon) at low pressure. A high voltage is applied to this chamber. The metallic chamber will conduct electricity. When radiation enters the chamber through the window, the photons in the radiation will ionize the inert gas inside the chamber. This will make the gas conductive. The electrons produced due to ionization are accelerated due to the potential that we applied and these electrons cause

even more ionization. The ionized electrons travel towards the anode. The anode is connected to a counter. The counter counts the electrons reaching the anode. This is how we measure radiation.

Resultant Curve: Geiger Plateau

The performance of a GM counter is typically illustrated by the Geiger Plateau curve, which shows the count rate versus the applied voltage.

1. **Initial Region:**
   At low voltages, no significant ionization occurs, so the count rate is very low.
2. **Threshold Region:**
   As voltage increases, the counter starts to detect ionizing events. The count rate begins to rise sharply.
3. **Geiger Plateau:**
   At a certain voltage range, the count rate levels off and remains relatively constant. This is the operational region of the GM counter. The plateau indicates that each ionizing event is creating a detectable pulse, and the counter is functioning reliably.
4. **Continuous Discharge Region:**
   If the voltage is increased too much, continuous discharge occurs, where the counter no longer operates correctly, and the count rate rises rapidly again.

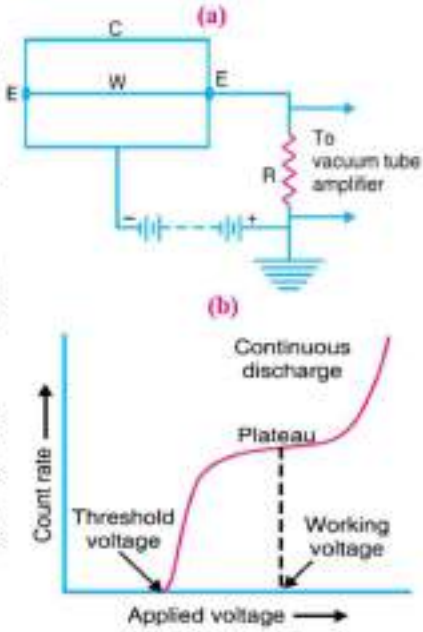


Figure 28.12 (a, b)
applied voltage vs counter rate graph

Use of GM Counter:

Geiger-Muller counters have a wide range of applications, including:

- **Radiation Monitoring:**
  They are commonly used for radiation monitoring in nuclear power plants, laboratories, and industrial settings to measure radiation levels and ensure the safety of workers and the environment."
- **Environmental Monitoring:**
  GM counters are employed for environmental radiation monitoring to assess background radiation levels and detect any abnormal increases in radiation.
- **Health Physics:**
  Health physicists use GM counters to monitor radiation exposure of individuals working with radioactive materials or in radiation-prone environments.
- **Education:**
  Health physicists use GM counters to monitor radiation exposure of individuals working with radioactive materials or in radiation-prone environments.
- **Radiological Emergencies:**
  In the event of radiological emergencies or accidents involving radioactive materials, GM counters can be used to assess radiation contamination levels.

EXERCISE

Section (A): Multiple Choice Questions (MCQs)
Choose the correct answer:

1. The Standard Model classifies elementary particles into two main groups:
   (a) Baryons and Leptons (b) Fermions and Bosons
   (c) Quarks and Gluons (d) Hadrons and Mesons
2. The following is NOT a flavor of quark:
   (a) Up (b) Down (c) Electron (d) Top
3. The charge of an up quark is:
   (a) +1/2e (b) -1/2 e (c) +2/3e (d) -2/3e
4. The concept of "quark confinement" implies that:
   (a) Quarks cannot exist as free particles outside of hadrons.
   (b) Quarks are always found in pairs with opposite charges.
   (c) Quarks have a strong affinity for gluons.
   (d) Quarks are the fundamental building blocks of all matter.
5. The primary role of the Higgs boson in the Standard Model is:
   (a) Mediating electromagnetic interactions (b) Providing mass to other particles
   (c) Transmitting the strong nuclear force (d) Creating dark matter
6. According to the Standard Model, the term "color" refers to:
   (a) Visible light spectrum (b) Charge property of quarks
   (c) Mass of particles (d) Spin of particles
7. A particle made up of a quark and an antiquark is called:
   (a) Lepton (b) Baryon
   (c) Meson (d) Neutrino
8. An elementary particle that feels all three fundamental forces (electromagnetic, weak, and strong nuclear forces) is:
   (a) Lepton (b) Quark
   (c) Electron (d) Neutrino
9. The primary function of a Geiger-Müller counter in particle physics is to:
   (a) Measure the velocity of particles
   (b) Detect and count ionizing radiation
   (c) Create antimatter particles
   (d) Generate magnetic fields
10. In a Geiger-Müller counter, the gas commonly used to detect ionizing radiation is:
    (a) Oxygen (b) Neon (c) Argon (d) Helium

Section (B): CRQs (Short Answered Questions):

1. Explain the difference between bosons and fermions and their roles in mediating fundamental forces.
2. Compare and contrast the properties of quarks and leptons, the two main categories of fermions in the Standard Model.
3. Define the term "lepton" and provide examples of leptons. Explain their fundamental properties and role in the Standard Model of particle physics.
4. Explain the concept of color charge in quarks and its significance in the strong nuclear force. How does the combination of quarks contribute to the color-neutral nature of protons and neutrons?
5. Describe the structure of a proton and neutron in terms of its quark composition. How do quarks combine to form a proton and a neutron, and what are the specific types of quarks involved?

Section (C): ERQs (Long Answered Questions):

1. Explain the structure of the Standard Model, including the different types of particles and their relationships. How does the model classify fundamental forces?
2. Describe in detail all fundamental forces and their associated field particles. What is a boson, and why are bosons referred to as field particles?
3. Describe the operating principle of a Geiger-Müller counter. How does it detect and quantify ionizing radiation, and what are its limitations in terms of measurement range and types of radiation detected?
4. Discuss the uses of a Wilson cloud chamber in particle physics experiments. How does it visualize charged particle tracks, and what factors can affect its performance?