Answers

1. **Answer: b**

Explanation:

Understanding Electrical Resistivity of Metals

Electrical resistivity is a fundamental property of a material that quantifies how strongly it opposes the flow of electric current. It is an intrinsic property, meaning it depends only on the material itself and its temperature, not on the size or shape of the object made from the material. The standard unit for electrical resistivity is the ohm-meter $(\Omega \cdot m)$.

Factors Affecting Electrical Resistivity

Several factors can influence the electrical resistivity of a metal:

- **Material Type:** Different metals have different atomic structures and electron configurations, which affects how easily electrons can move through them.
- **Temperature:** For most metals, resistivity increases with increasing temperature. This is because increased thermal vibrations of atoms scatter the electrons more, impeding their flow.
- **Impurities and Defects:** The presence of impurities or structural defects in the metal lattice can also scatter electrons and increase resistivity.

Typical Electrical Resistivity Values

To determine the correct order of electrical resistivity for Silver (Ag), Tungsten (W), Nickel (Ni), and Mercury (Hg), we need to look at their typical resistivity values at a standard temperature, like room temperature $(20°C$ to $25°C)$. Note that these values are approximate and can vary slightly depending on the purity and specific conditions.

Approximate Electrical Resistivity at $20°C$

Metal	Symbol	Resistivity $(\times 10^{-8}\Omega \cdot m)$
Silver	Ag	1.59
Tungsten	W	5.60
Nickel	Ni	6.99
Mercury	Hg	98.00

Comparing Resistivity Values

Let's compare the resistivity values from the table:

- Mercury (Hg): $98.00 \times 10^{-8}\Omega \cdot m$
- Nickel (Ni): $6.99 \times 10^{-8}\Omega \cdot m$
- Tungsten (W): $5.60 \times 10^{-8}\Omega \cdot m$
- Silver (Ag): $1.59 \times 10^{-8}\Omega \cdot m$

Arranging these values in decreasing order of resistivity, we get:

$98.00 > 6.99 > 5.60 > 1.59$

This corresponds to the order of metals:

Mercury > Nickel > Tungsten > Silver

In terms of their chemical symbols, the order of electrical resistivity is:

$Hg > Ni > W > Ag$

Silver has the lowest electrical resistivity among common metals, making it an excellent conductor. Mercury, being a liquid metal at room temperature, has a significantly higher resistivity compared to the solid metals listed.

Therefore, the correct order of electrical resistivity from highest to lowest is Mercury, Nickel, Tungsten, and Silver.

Revision Table: Metal Resistivity Order

Electrical Resistivity Comparison

Metal	Symbol	Approximate Resistivity ($\times 10^{-8} \Omega \cdot m$)
Mercury	Hg	98.00
Nickel	Ni	6.99
Tungsten	W	5.60
Silver	Ag	1.59

Additional Information on Electrical Properties

Understanding electrical resistivity is crucial in selecting materials for various applications. Related concepts include:

- **Electrical Conductivity:** This is the inverse of resistivity ($\sigma = 1/\rho$) and measures how easily electric current flows through a material. Materials with low resistivity have high conductivity. Silver has the highest electrical conductivity among all metals.
- **Applications:** Metals with low resistivity like copper and aluminum are used extensively in electrical wiring. Tungsten's high melting point and relatively higher resistivity make it suitable for filaments in incandescent light bulbs. Nickel is used in alloys and battery electrodes. Mercury's conductivity makes it useful in some switches and older thermometers (though its use is declining due to toxicity).
- **Temperature Dependence:** For metals, resistivity generally increases linearly with temperature over a certain range. This property is used in resistance thermometers.

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2. **Answer: d**

Explanation:

Understanding Blood Sugar Detection by the Pancreas

The question asks which organ is responsible for detecting a rise in blood sugar levels. Let's look at the roles of the options provided.

- **Liver:** The liver plays a crucial role in regulating blood sugar by storing and releasing glucose, but it's not the primary organ that *detects* the initial rise and signals for insulin release in response to a meal.
- **Gallbladder:** The gallbladder stores bile produced by the liver, which helps in digesting fats. It has no direct role in detecting blood sugar levels.
- **Kidney:** The kidneys filter blood and can remove excess glucose from the body when blood sugar levels are very high (a condition called glycosuria), but they don't detect the initial rise that triggers metabolic regulation.
- **Pancreas:** The pancreas is a key organ in blood sugar regulation. It contains clusters of endocrine cells called the islets of Langerhans. Within these islets, specific cells called beta cells are highly sensitive to changes in blood glucose levels. When blood sugar rises (for example, after eating), the beta cells of the pancreas detect this increase and release insulin, a hormone that helps lower blood sugar.

Therefore, the pancreas, specifically the beta cells in its islets of Langerhans, is the organ that detects the rise in blood sugar levels.

Organ	Primary Role in Blood Sugar	Detects Rise?
Liver	Stores/Releases Glucose	No (Primary Detector)
Gallbladder	Stores Bile (Fat Digestion)	No
Kidney	Filters Blood, Excretes Excess (High Levels)	No (Primary Detector)
Pancreas	Produces Insulin/Glucagon	Yes

Based on the function of these organs in the human body, the pancreas is the correct answer as it directly monitors and responds to changes in blood glucose concentration.

Revision Table: Key Organs in Blood Sugar Regulation

Organ	Hormone Produced / Action	Effect on Blood Sugar
Pancreas (Beta cells)	Insulin	Lowers blood sugar (promotes glucose uptake/storage)
Pancreas (Alpha cells)	Glucagon	Raises blood sugar (promotes glucose release from liver)
Liver	Stores glucose as glycogen; releases glucose	Regulates blood sugar levels
Kidneys	Filters blood	Removes excess glucose if levels are very high

Additional Information on Blood Sugar Detection and Response

The detection of blood sugar levels by the pancreas is a critical part of the body's homeostasis. Homeostasis is the process by which the body maintains a stable internal environment. When blood glucose levels rise above a certain point, the beta cells in the pancreas increase insulin secretion. Insulin acts on various cells (like muscle and fat cells) to help them take up glucose from the blood, and it also promotes the liver and muscles to store glucose as glycogen. This helps to bring the elevated blood sugar levels back down to the normal range.

Conversely, when blood sugar levels fall too low, the alpha cells in the pancreas detect this drop and release glucagon. Glucagon primarily acts on the liver, stimulating it to break down stored glycogen into glucose and release it into the bloodstream, thereby raising blood sugar levels.

This constant interplay between insulin and glucagon, regulated by the pancreas's detection of blood glucose levels, is vital for maintaining energy balance and preventing conditions like hyperglycemia (high blood sugar) or hypoglycemia (low blood sugar), which are associated with diseases like diabetes mellitus.

3. **Answer: a**

Explanation:

Solving the Blood Relation Puzzle

Let's break down the relationships described in the question step by step to determine how F is related to E. This is a common type of blood relation question found in reasoning sections of exams.

Understanding the Given Relations

We are given the following facts:

- F is the father of S and D. This means S and D are siblings, and F is their parent.
- D's Paternal aunt's daughter is E. Let's analyze this part carefully.

Step-by-Step Relationship Deduction

To find the relationship between F and E, we need to trace the connection starting from D's paternal aunt:

1. **D's father:** The question states F is the father of D. So, D's father is F.
2. **D's Paternal aunt:** A paternal aunt is the sister of one's father. So, D's paternal aunt is the sister of D's father, who is F. This means D's paternal aunt is F's sister.
3. **D's Paternal aunt's daughter:** This is the daughter of F's sister.
4. **E:** The question states that D's paternal aunt's daughter is E. Therefore, E is the daughter of F's sister.

Now we know that E is the daughter of F's sister. What does this make F in relation to E?

If F is the brother of E's mother, then F is E's maternal uncle.

Identifying the Relationship between F and E

Based on our deduction:

- E's mother is F's sister.
- F is the brother of E's mother.

This establishes that F is the maternal uncle of E.

Reviewing the Options

Let's look at the provided options:

1. Maternal Uncle
2. Daughter-in-law
3. Daughter
4. Nephew

Our analysis concluded that F is E's maternal uncle, which matches Option 1.

Final Answer Explanation

The question states F is the father of D. D's paternal aunt is her father's sister, which means F's sister. E is the daughter of F's sister. Therefore, F is the maternal uncle of E.

Person	Relationship
F	Father of D
D’s Paternal Aunt	F's Sister
E	Daughter of F's Sister
F	E's Maternal Uncle (Brother of E's Mother)

Revision Table: Blood Relation Terms

Relationship	Description
Paternal Aunt	Father's Sister
Maternal Aunt	Mother's Sister
Paternal Uncle	Father's Brother
Maternal Uncle	Mother's Brother
Nephew	Brother's or Sister's Son
Niece	Brother's or Sister's Daughter

Additional Information: Solving Blood Relation Questions

Blood relation questions test your ability to understand complex family relationships. Here are some tips for solving them effectively:

- Draw a family tree: Visual representation can make the relationships much clearer. Use symbols for gender and connections.
- Break down the statement: Analyze the given information part by part, starting from the person mentioned first or last in a chain of relations.
- Substitute relations: Replace complex phrases like "paternal aunt's daughter" with simpler ones like "cousin" or "sister's daughter" as you deduce them.
- Be careful with gender: Pay close attention to whether a person is male or female if mentioned.
- Practice: Solving various types of blood relation puzzles will improve your speed and accuracy.

In this specific question, identifying "D’s Paternal aunt" as "F’s sister" was the key step to correctly determining the relationship between F and E.

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4. **Answer: c**

Explanation:

Identifying the Author of 'A Book of Light'

The question asks us to identify the author of the book titled ‘A Book of Light: When a Loved One Has a Different Mind’. This book is a significant work known for its exploration of complex themes.

Let's look at the options provided:

- Arundhati Roy
- Amish Tripathi
- Jerry Pinto
- Harper Lee

Each of these individuals is a well-known author, but their works span different genres and subjects.

Upon reviewing the authorship of prominent books, we find that ‘A Book of Light: When a Loved One Has a Different Mind’ was written by Jerry Pinto.

About the Book and Author

‘A Book of Light: When a Loved One Has a Different Mind’ is a non-fiction book compiled and edited by Jerry Pinto. It features various contributors sharing their experiences and perspectives on living with and caring for loved ones who have mental health conditions or significant cognitive differences. The book provides personal accounts, reflections, and insights into the challenges and realities faced by families.

Jerry Pinto is an Indian author, poet, translator, and journalist. He has written several acclaimed books, including the novel ‘Em and the Big Hoom’, which won the Sahitya Akademi Award. His work often delves into sensitive and complex human experiences. ‘A Book of Light’ is one of his notable contributions to literature, offering a platform for important conversations about mental health and family dynamics.

Comparing this information with the options, it becomes clear that Jerry Pinto is the author/editor of the specified book.

Therefore, the correct answer is Jerry Pinto.

Revision Table: Key Authors and Works

Author	Notable Works	Genre/Focus
Arundhati Roy	The God of Small Things, The Ministry of Utmost Happiness	Fiction, Essays (Political & Social Commentary)
Amish Tripathi	The Shiva Trilogy, Ram Chandra series	Mythological Fiction
Jerry Pinto	Em and the Big Hoom, A Book of Light, Island of Lost Shadows	Fiction, Non-fiction (Essays, Anthology), Poetry, Translation
Harper Lee	To Kill a Mockingbird, Go Set a Watchman	Fiction (American Southern Literature)

Additional Information: Understanding 'A Book of Light'

'A Book of Light' is considered an important collection because it brings together diverse voices to talk about a topic often surrounded by stigma – mental illness and cognitive differences within families. The essays cover a range of conditions and experiences, providing a nuanced look at the emotional, practical, and social aspects of caregiving and relationships in these circumstances. Jerry Pinto's role as editor was crucial in curating these powerful narratives, making the book a valuable resource for understanding and empathy.

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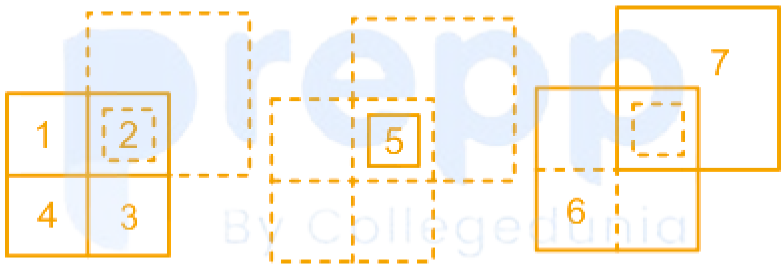
5. **Answer: c**

Explanation:

The number of squares is as follows:



Hence, there are seven squares in the given figure.

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6. **Answer: a**

Explanation:

Finding the Length of a Rectangle from its Perimeter and Ratio

This problem involves finding the dimensions of a rectangle when we know the ratio of its length and breadth and its total perimeter. We'll use the given ratio to express the length and breadth in terms of a variable and then apply the formula for the perimeter of a rectangle.

Understanding the Problem

- The length and breadth of the rectangle are in the ratio 3 : 1. This means for every 3 units of length, there is 1 unit of breadth.
- The perimeter of the rectangle is 96 m.
- We need to find the actual length of the rectangle.

Setting up the Dimensions with the Ratio

Since the ratio of length to breadth is 3 : 1, we can represent the length and breadth using a common multiple, let's call it 'x'.

- Let the length (l) be $3x$.
- Let the breadth (b) be $1x$, or simply x.

Here, 'x' is a positive value representing a unit of measurement relative to the ratio.

Using the Perimeter Formula

The formula for the perimeter (P) of a rectangle is given by:

$P = 2 \times (\text{length} + \text{breadth})$

Substituting the expressions for length and breadth $(l = 3x, b = x)$ and the given perimeter $(P = 96 \text{ m})$ into the formula:

$96 = 2 \times (3x + x)$

$96 = 2 \times (4x)$

$96 = 8x$

Solving for the Variable 'x'

To find the value of 'x', we need to isolate 'x' in the equation $96 = 8x$. We can do this by dividing both sides of the equation by 8:

$\frac{96}{8} = \frac{8x}{8}$

$12 = x$

So, the value of the variable 'x' is 12.

Calculating the Length of the Rectangle

Now that we know $x = 12$, we can find the length of the rectangle using our expression for length, which was $3x$.

$\text{Length} = 3x$

$\text{Length} = 3 \times 12$

$\text{Length} = 36 \text{ m}$

Let's also find the breadth for completeness:

$\text{Breadth} = x$

$\text{Breadth} = 12 \text{ m}$

Verifying the Solution

We can check our answer by calculating the perimeter using the found length and breadth:

$P = 2 \times (\text{length} + \text{breadth})$

$P = 2 \times (36 + 12)$

$P = 2 \times (48)$

$P = 96 \text{ m}$

This matches the given perimeter, confirming our calculations are correct.

Conclusion

The length of the rectangle is 36 m.

Revision Table: Rectangle Dimensions and Perimeter

Concept	Description	Formula/Relation
Rectangle	A quadrilateral with four right angles. Opposite sides are equal in length.	
Length (l)	The longer side of the rectangle (or the side referred to as length).	
Breadth (b)	The shorter side of the rectangle (or the side referred to as breadth).	
Ratio of l : b	Expresses the relationship between the lengths of the sides. If ratio is a:b, l=ax, b=bx for some x.	$l : b = 3 : 1$ in this problem
Perimeter (P)	The total distance around the boundary of the rectangle.	$P = 2(l + b)$

Additional Information: Solving Ratio Problems

When a ratio is given between two quantities, like length and breadth in this case, representing them using a variable is a common and effective technique. Here's a breakdown:

- **Identify the Ratio:** Note the given ratio, e.g., a : b.
- **Introduce a Variable:** Let a common variable (say, x) multiply each part of the ratio. The quantities then become ax and bx.
- **Formulate an Equation:** Use the other information given in the problem (like perimeter, area, sum, difference) to create an equation involving the variable x.
- **Solve the Equation:** Find the value of x.
- **Calculate the Quantities:** Substitute the value of x back into the expressions (ax and bx) to find the actual values of the quantities.

This method is applicable to various problems involving ratios in geometry, algebra, and other areas.

7. **Answer: c**

Explanation:

Understanding Group 18 in the Modern Periodic Table

The question asks about the number of elements found in the 18th group of the Modern Periodic Table. The Modern Periodic Table organizes elements based on their atomic number and recurring chemical properties. Vertical columns in the periodic table are called groups, and horizontal rows are called periods.

Identifying Group 18 Elements

Group 18 is the rightmost group in the Modern Periodic Table. It is also known as the Noble Gases or inert gases because of their generally low reactivity. The elements in this group are:

- Helium (He)
- Neon (Ne)
- Argon (Ar)
- Krypton (Kr)
- Xenon (Xe)
- Radon (Rn)
- Oganesson (Og)

Each of these elements occupies a specific period in Group 18.

Counting the Elements in Group 18

Let's list the elements along with the period they belong to:

Element	Symbol	Period
Helium	He	1
Neon	Ne	2
Argon	Ar	3
Krypton	Kr	4
Xenon	Xe	5
Radon	Rn	6
Oganesson	Og	7

By counting the elements listed in the table and the bullet points, we find there are a total of 7 elements in Group 18 of the Modern Periodic Table.

Conclusion on Group 18 Elements

Based on the composition of the Modern Periodic Table, Group 18 contains Helium (He), Neon (Ne), Argon (Ar), Krypton (Kr), Xenon (Xe), Radon (Rn), and Oganesson (Og). This gives a total of 7 elements.

Revision Table: Key Facts about Group 18

Group Number	Common Name	Number of Elements	Typical Reactivity
18	Noble Gases	7	Very low (inert)

Additional Information: Periodic Table Structure

The Modern Periodic Table is structured into 7 periods and 18 groups. Groups contain elements with similar chemical properties due to having the same number of valence electrons (except for Helium). Periods represent the principal energy level of the valence electrons. The number of elements in each period varies:

- Period 1: 2 elements (H, He)
- Period 2: 8 elements
- Period 3: 8 elements
- Period 4: 18 elements
- Period 5: 18 elements
- Period 6: 32 elements (including Lanthanides)
- Period 7: 32 elements (including Actinides, though some might be undiscovered or synthetic)

Group 18 has one element from each of the 7 periods, resulting in a total of 7 elements.

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8. Answer: c

Explanation:

- Xenon's name comes from the Greek word Xenos, which means ' **stranger** '.
- Xenon is an element belonging to the group of **noble gases** i.e. **18th group** .
- Xenon is used in photographic flashes, in high-pressure arc lamps for motion picture projection, and in high-pressure arc lamps to produce ultraviolet light.
- Xenon is a chemical element with symbol **Xe** and atomic number 54.
- It was discovered by **William Ramsay and Morris Travers** in **1898.**
- It is a rare, colourless and odourless heavy gas.
- It is used in instruments for **radiation detection** .

Note:

Neutrons, Protons and Electrons are the subatomic particle of an atom

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9. **Answer: a**

Explanation:

Analysing the Statement and Assumptions on Discipline Issues

Let's break down the given statement and assumptions to determine which one is implicit. The statement is:

"The headmistress called for an urgent meeting for all staff members to discuss on discipline issues."

An assumption is something taken for granted or accepted as true without proof. We need to see what must be true for the statement to make sense.

Statement Analysis

The statement tells us several key things:

- A specific person, the headmistress, took an action.
- The action was calling a meeting.
- The meeting was urgent.
- The meeting was for all staff members.
- The purpose of the meeting was to discuss discipline issues.

The fact that a meeting, especially an urgent one, has been called specifically to discuss "discipline issues" strongly suggests that such issues exist or have recently occurred.

Assumption I: There were some discipline issues that were raised earlier.

Let's consider if this assumption must be true. If there were absolutely no discipline issues (current or past), why would the headmistress call an urgent meeting *specifically* to discuss them? The term "discipline issues" implies that problems related to discipline are present or have been noted. While the assumption uses the phrase "raised earlier," the core idea that "discipline issues" exist is essential for the meeting's purpose. It is highly implicit that there are existing or recent discipline problems to warrant an urgent discussion among all staff. This assumption is necessary for the statement to be logical.

Assumption II: The headmistress likes to address all of them together.

This assumption talks about the headmistress's preference or general method of handling discipline issues. The statement only tells us that *this specific* urgent meeting is being held with all staff to discuss these issues. This action shows *one way* she is addressing the issues in this instance. However, it does not provide any information about whether this is her preferred method for *all* issues, or if she *likes* this method in general. She might be doing it this way due to the urgency or the nature of the issues, not necessarily because it's her favourite approach. Therefore, this assumption is not implicitly true based solely on the statement.

Conclusion

Based on the analysis:

- Assumption I: It is necessary that discipline issues exist to justify an urgent meeting to discuss them. Thus, Assumption I is implicit.
- Assumption II: The statement only describes one instance of how issues are being addressed (a meeting with all staff). It does not imply the headmistress's general preference or that she likes to address all issues this way. Thus, Assumption II is not implicit.

Therefore, only Assumption I is implicit in the statement.

Component	Analysis	Implicit?	
Statement Purpose	Urgent meeting to discuss discipline issues.	Suggests issues exist.	
Assumption I	Discipline issues were raised earlier (or exist).	Necessary for the meeting's purpose.	**Yes**
Assumption II	Headmistress likes addressing all issues together.	Describes preference, not a necessary condition for this meeting.	No

Revision Table: Statement and Assumption Concepts

Term	Definition/Explanation
Statement	A declarative sentence or assertion that can be judged as true or false.
Assumption	Something taken for granted or accepted as true without proof; an underlying belief or premise.
Implicit Assumption	An assumption that is not stated directly but is understood to be true based on the statement made. It is a necessary condition for the statement to be meaningful or logical.
Explicit Assumption	An assumption that is directly stated in the text or argument.

Additional Information: Identifying Implicit Assumptions

When analysing statements for implicit assumptions, consider the following:

- **What must be true for the statement to make sense?** Think about the conditions that enable the action or situation described in the statement.

- **What is the purpose of the statement?** The reason behind the statement can reveal underlying assumptions.
- **What is the context?** While the question provides a statement in isolation, consider the typical context in which such a statement would be made (e.g., a school environment for a headmistress).
- **Eliminate alternatives:** If the assumption were false, would the statement still be logical or possible? If not, the assumption is likely implicit.
- Be careful not to confuse implicit assumptions with logical inferences or conclusions drawn *from* the statement. An assumption is something that must be true *before* the statement is made or for it to be valid.

In this question, the statement about discussing "discipline issues" inherently assumes the existence of such issues. The urgency and the involvement of "all staff" are details about *how* the existing or recent issues are being addressed, but the core assumption is that the issues themselves are real.

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10. **Answer: b**

Explanation:

Understanding Work Done Against Gravity

Work done is a physics concept that describes the energy transferred when a force moves an object over a distance. When you lift an object upwards against gravity, you are doing work on the object. The work done against gravity depends on the object's mass, the gravitational acceleration, and the height the object is lifted.

Calculating Work Done in Lifting Luggage

The work done (W) when lifting an object of mass (m) vertically through a height (h) against gravity (g) can be calculated using the formula:

$W = \text{Force} \times \text{Distance}$

In this case, the force you apply to lift the luggage is equal to the weight of the luggage, which is $F = m \times g$. The distance is the height (h) the luggage is lifted. So, the work done against gravity is:

$$W = m \times g \times h$$

Given Information

Let's identify the values given in the question:

- Mass of the luggage (m) = 20 kg
- Height lifted (h) = 2 m
- Acceleration due to gravity (g) = 10 ms^{-2}

Step-by-Step Work Done Calculation

Now, we can plug these values into the work done formula:

$$W = m \times g \times h$$

$$W = 20\,\text{kg} \times 10\,\text{ms}^{-2} \times 2\,\text{m}$$

Multiplying these values:

$$W = 200\,\text{kg} \cdot \text{m/s}^2 \times 2\,\text{m}$$

$$W = 400\,\text{kg} \cdot \text{m}^2/\text{s}^2$$

The unit kg·m^2/s^2 is equivalent to Joules (J), which is the standard unit for work and energy.

$$W = 400\,\text{J}$$

Final Answer

The work done by the person to lift the luggage is 400 Joules.

Work Done Calculation Revision Table

Quantity	Symbol	Value	Unit
Mass	m	20	kg
Gravity	g	10	ms^{-2}
Height	h	2	m
Work Done	W	?	J

Formula used: $W = mgh$

Additional Information on Work and Energy

- Work done is a scalar quantity, meaning it only has magnitude and no direction.
- When work is done against gravity, the potential energy of the object increases. The increase in potential energy is equal to the work done, assuming no energy is lost to friction or air resistance.
- The standard unit for work and energy in the International System of Units (SI) is the Joule (J). One Joule is defined as the work done when a force of one Newton moves an object one meter in the direction of the force.
- Lifting an object at a constant velocity means the net force is zero, but work is still done by the lifting force against the force of gravity.

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11. **Answer: a**

Explanation:

Understanding the Problem: Finding the Smallest Natural Number

The question asks for the smallest natural number that, when divided by several different numbers (9, 10, 12, and 15), always leaves the same remainder, which is 3. A natural number is a positive whole number (1, 2, 3, ...).

To solve this kind of problem, we need to find a number that is a multiple of the divisors (9, 10, 12, 15) plus the constant remainder (3).

Let the required number be N. According to the problem:

- $N \div 9$ leaves a remainder of 3.
- $N \div 10$ leaves a remainder of 3.
- $N \div 12$ leaves a remainder of 3.
- $N \div 15$ leaves a remainder of 3.

This means that if we subtract 3 from N, the resulting number $(N-3)$ must be perfectly divisible by 9, 10, 12, and 15. In other words, $(N-3)$ must be a common multiple of 9, 10, 12, and 15.

Since we are looking for the **smallest** such number N, the smallest possible value for $(N-3)$ will be the Least Common Multiple (LCM) of 9, 10, 12, and 15.

Calculating the LCM of 9, 10, 12, and 15

To find the LCM, we first find the prime factorization of each number:

- Prime factors of 9: $9 = 3 \times 3 = 3^2$
- Prime factors of 10: $10 = 2 \times 5$
- Prime factors of 12: $12 = 2 \times 2 \times 3 = 2^2 \times 3$
- Prime factors of 15: $15 = 3 \times 5$

Now, we take the highest power of each prime factor that appears in any of the factorizations:

- Highest power of 2 is 2^2.
- Highest power of 3 is 3^2.
- Highest power of 5 is 5^1.

The LCM is the product of these highest powers:

$$\text{LCM}(9, 10, 12, 15) = 2^2 \times 3^2 \times 5^1 = 4 \times 9 \times 5 = 36 \times 5 = 180$$

So, the smallest number that is perfectly divisible by 9, 10, 12, and 15 is 180.

Finding the Smallest Natural Number with Remainder 3

We know that $(N-3)$ is a multiple of the LCM (180). The smallest possible value for $(N-3)$ is the LCM itself, which is 180.

So, $N-3=180$.

To find N, we add 3 to the LCM:

$N=180+3=183$

Thus, the smallest natural number that leaves a remainder of 3 when divided by 9, 10, 12, or 15 is 183.

Verification

Let's check if 183 leaves a remainder of 3 when divided by 9, 10, 12, and 15:

- $183 \div 9$: $183=9 \times 20+3$. Remainder is 3.
- $183 \div 10$: $183=10 \times 18+3$. Remainder is 3.
- $183 \div 12$: $183=12 \times 15+3$. Remainder is 3.
- $183 \div 15$: $183=15 \times 12+3$. Remainder is 3.

The number 183 satisfies the conditions. Since we used the smallest common multiple (LCM), 183 is the smallest such natural number.

Exploring the Options

Let's quickly look at the other options to confirm they don't fit the criteria for the smallest number:

Option	Division by 9 (Remainder)	Division by 10 (Remainder)	Division by 12 (Remainder)	Division by 15 (Remainder)	Satisfies Condition?
183	3	3	3	3	Yes
153	0	3	9	3	No
63	0	3	3	3	No
123	6	3	3	3	No

As shown in the table, only 183 leaves a remainder of 3 in all cases. Therefore, 183 is indeed the smallest such number.

Revision Table: Smallest Natural Number Problem

Concept	Description	Application Here
Natural Numbers	Positive whole numbers (1, 2, 3, ...)	The required number must be a natural number.
Remainder	The amount left over after division	The problem specifies a constant remainder of 3.
Divisibility	When a number divides another with no remainder	$N - 3$ is divisible by 9, 10, 12, and 15.
Least Common Multiple (LCM)	The smallest positive integer that is a multiple of two or more numbers	Used to find the smallest number divisible by all divisors (9, 10, 12, 15).

Additional Information: Remainder Theorem and LCM

This type of problem is a classic application of the concept of LCM. When a number N leaves the same remainder r when divided by several numbers $d_1, d_2, ..., d_k$, it means that $(N - r)$ is perfectly divisible by each of $d_1, d_2, ..., d_k$. Therefore, $(N - r)$ must be a common multiple of $d_1, d_2, ..., d_k$.

To find the smallest such number N, we need the smallest possible value for $(N - r)$, which is the Least Common Multiple (LCM) of $d_1, d_2, ..., d_k$.

So, the smallest number N is given by: $N = \text{LCM}(d_1, d_2, ..., d_k) + r$.

In our case, the divisors are 9, 10, 12, and 15, and the remainder is 3. The smallest natural number is $\text{LCM}(9, 10, 12, 15) + 3 = 180 + 3 = 183$.

This principle holds true as long as the remainder is less than all the divisors.

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12. **Answer: a**

Explanation:

Understanding the Location of Rama Setu

Rama Setu, also widely known as Adam's Bridge, is a chain of limestone shoals, between Pamban Island, off the south-eastern coast of Tamil Nadu, India, and Mannar Island, off the north-western coast of Sri Lanka.

Where is Rama Setu Located Geographically?

To understand the location of Rama Setu, we need to look at the geography of the region connecting India and Sri Lanka. This area includes important water bodies and landforms.

- The chain of shoals runs between the Palk Strait in the north and the Gulf of Mannar in the south.
- The Palk Strait is a strait between the Tamil Nadu state of India and the Jaffna District of the Northern Province of the island nation of Sri Lanka.
- It connects the Palk Bay in the southwest with the Bay of Bengal in the northeast.

Considering this geographical context, the Rama Setu is located within this specific region that separates the Palk Strait from the Gulf of Mannar.

Analyzing the Options for Rama Setu's Location

Let's examine the given options:

1. **Palk Strait:** As discussed, the Palk Strait lies to the north of the Rama Setu. The Rama Setu effectively acts as a bridge separating the Palk Strait from the Gulf of Mannar. Therefore, stating that Rama Setu is located in the Palk Strait area or separating the Palk Strait from the Gulf of Mannar is accurate in geographical terms.
2. **Strait of Gibraltar:** The Strait of Gibraltar is a narrow strait that connects the Atlantic Ocean to the Mediterranean Sea and separates the Iberian Peninsula in Europe from Morocco in Africa. This location is geographically very distant from India and Sri Lanka.
3. **Kiel Canal:** The Kiel Canal is a 98-kilometre-long freshwater canal in the German state of Schleswig-Holstein. It connects the North Sea at Brunsbüttel to the Baltic Sea at Kiel-Holtenau. A canal is a man-made waterway, not a natural strait or feature like Rama Setu, and it is located in Europe, far from India.
4. **Bering Strait:** The Bering Strait is a strait of the Pacific Ocean, which separates Russia and the United States. It is located between Chukotka Peninsula of Russia in the west and the Seward Peninsula of Alaska in the east. This is also a location thousands of kilometers away from India and Sri Lanka.

Based on the geographical location of Rama Setu between India and Sri Lanka, and considering the Palk Strait is the major water body immediately north of this feature, the correct location among the given options is the Palk Strait region.

Revision Table: Key Straits and Canals

Geographical Feature	Location	Connects
Rama Setu (Adam's Bridge)	Between India & Sri Lanka	Separates Palk Strait (North) from Gulf of Mannar (South)
Palk Strait	Between India & Sri Lanka	Palk Bay & Bay of Bengal
Strait of Gibraltar	Between Europe & Africa	Atlantic Ocean & Mediterranean Sea
Kiel Canal	Germany	North Sea & Baltic Sea (Man-made)
Bering Strait	Between Asia & North America	Pacific Ocean & Arctic Ocean

Additional Information on Rama Setu and Palk Strait

- Rama Setu is sometimes referred to as a "bridge" due to its appearance from satellite images and historical/mythological accounts.
- The Palk Strait is relatively shallow, which makes navigation difficult for large ships.
- The geographical features in this region, including the Palk Strait and the Gulf of Mannar, are significant for biodiversity.
- Adam's Bridge is also important from a geological perspective, with debates about its formation.

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13. **Answer: b**

Explanation:

Understanding the Role of the Placenta in Human Pregnancy

The placenta is a temporary organ that develops in the uterus during pregnancy. It provides oxygen and nutrients to the growing baby and removes waste products from the baby's blood. It attaches to the wall of the uterus, and the baby's umbilical cord arises from it.

Key Functions of the Placenta

The placenta performs several vital functions for the developing embryo/fetus:

- **Nutrient Supply:** Essential nutrients like glucose, amino acids, fatty acids, vitamins, and minerals are transferred from the mother's blood to the baby's blood across the placenta.
- **Oxygen Supply:** Oxygen from the mother's blood diffuses across the placenta into the baby's blood.
- **Waste Removal:** Waste products produced by the baby, such as carbon dioxide and urea, diffuse from the baby's blood across the placenta into the mother's blood, which are then filtered and removed by the mother's body.
- **Hormone Production:** The placenta produces important hormones like human chorionic gonadotropin (hCG), progesterone, and estrogen, which are crucial for maintaining the pregnancy.
- **Immune Protection:** Antibodies from the mother's blood can pass across the placenta, providing passive immunity to the baby during pregnancy and for a short period after birth.
- **Support and Attachment:** The placenta provides physical support by anchoring the developing embryo/fetus within the uterus.

Analyzing the Given Options

Let's evaluate each option based on the known functions of the placenta:

- **Option 1: Support the embryo** - While the placenta does provide structural support by attaching the embryo to the uterine wall, this is just one aspect and not its primary, most defining function compared to nutrient/waste exchange.
- **Option 2: Remove waste matter from the embryo and provide nutrition to the embryo** - This option accurately combines the two major life-sustaining

functions of the placenta: providing necessary substances (nutrition) and removing harmful ones (waste).

- **Option 3: Provide nutrition to the embryo** - This is a correct function, but it is only one part of the complete picture. The placenta also removes waste.
- **Option 4: Remove waste matter from the embryo** - This is also a correct function, but again, it is only one part. The placenta also provides nutrition.

Comparing the options, Option 2 provides the most comprehensive description of the essential physiological work performed by the placenta tissue for the embryo.

Placenta Functions Summary

Function	Description
Nutrition Supply	Transfers nutrients from mother to embryo.
Waste Removal	Transfers waste from embryo to mother.
Oxygen Supply	Transfers oxygen from mother to embryo.
Hormone Production	Produces hormones essential for pregnancy.
Immune Transfer	Passes antibodies from mother to embryo.
Support	Anchors the embryo to the uterine wall.

Therefore, the primary work done by the placenta tissue in humans is facilitating the exchange of substances between the mother and the embryo, specifically providing nutrition and removing waste matter.

Revision Table: Placenta's Vital Roles

Key Role	Explanation
Nutrition & Oxygen	Supplies the embryo with everything it needs to grow.
Waste & CO_2 Removal	Takes away the embryo's waste products.
Hormone Production	Helps maintain the pregnancy.
Protection	Acts as a barrier and transfers antibodies.

Additional Information: Placenta Structure and Importance

The placenta is a complex organ with a rich blood supply from both the mother and the embryo. Maternal blood and fetal blood do not mix directly but are separated by a thin barrier called the placental barrier. This barrier allows for the selective passage of substances. The placenta is crucial for the healthy development and survival of the embryo and fetus throughout the pregnancy. Without a functional placenta, the embryo would not receive the necessary resources for growth and development and would be unable to eliminate metabolic wastes.

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14. **Answer: c**

Explanation:

Understanding Blood Relations

Blood relation questions are common in various exams. They test your ability to understand family relationships based on given information. The key is to break down the relationships step by step and identify the connection between the individuals mentioned.

Analyzing the Spandana and Raju Relationship

The question states: "Spandana is the sister of Raju."

This simple statement establishes a direct sibling relationship. Spandana and Raju share the same parents.

Tracing the Relation: Raju's Father's Sister

Now, let's look at the second part of the question: "How is Raju’s father’s sister related to Spandana?"

We need to trace the relationship from Raju to his father's sister and then determine that person's relationship to Spandana.

1. Start with Raju.
2. Go to Raju's father.
3. Then, consider Raju's father's sister.

Since Spandana is Raju's sister, Raju's father is also Spandana's father.

Therefore, Raju's father's sister is also Spandana's father's sister.

Identifying the Relationship to Spandana

What do we call our father's sister?

- Your father's sister is your paternal aunt.
- Your mother's sister is your maternal aunt.

In this case, Raju's father's sister is Spandana's father's sister, which means she is Spandana's paternal aunt.

The term "aunt" refers to both paternal and maternal aunts.

Conclusion on the Relationship

Based on the analysis:

- Spandana and Raju are siblings.
- Raju's father is Spandana's father.
- Raju's father's sister is Spandana's father's sister.

- A father's sister is an aunt.

Thus, Raju's father's sister is related to Spandana as her Aunt.

Individual A	Individual B	Relation
Spandana	Raju	Siblings (Sister/Brother)
Raju	Raju's Father	Son/Father
Spandana	Spandana's Father	Daughter/Father
Raju's Father's Sister	Raju's Father	Sister/Brother
Raju's Father's Sister	Spandana's Father	Sister/Brother
Raju's Father's Sister	Spandana	Aunt/Niece

Revision Table: Common Blood Relations

Relationship Description	Term
Father's sister	Paternal Aunt
Mother's sister	Maternal Aunt
Father's brother	Paternal Uncle
Mother's brother	Maternal Uncle
Father's father	Paternal Grandfather
Mother's mother	Maternal Grandmother
Son's son	Grandson
Daughter's daughter	Granddaughter

Additional Information on Blood Relations Puzzles

When solving blood relation puzzles, it's helpful to visualize the relationships or draw a family tree. Start with the people mentioned first and build connections based on the information provided. Pay close attention to gender if specified, although in this problem, the gender of Raju's father's sister is clear (sister implies female), and Spandana's gender is also clear (sister implies female). The gender of Raju doesn't impact the relationship of his father's sister to Spandana.

Common relationships involved include parents, siblings, children, grandchildren, grandparents, uncles, aunts, nephews, nieces, cousins, in-laws, etc.

Practice with various types of blood relation problems helps in quickly identifying the connections.

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15. **Answer: c**

Explanation:

Analyse the Number Series Pattern

The question presents a series of elements, each structured in the format: First Number – Second Number – 96. The number 96 is constant across all elements in the series. We need to find a rule or pattern that applies to most elements in the series and identify which element among the given options does not follow this pattern.

Understanding the Series Structure

The series is:

- 5 – 1 – 96
- 27 – 1 – 96
- 18 – 2 – 96
- 12 – 3 – 96
- 2 – 4 – 96

Let's denote the first number as A, the second number as B, and the third number as C. So, each element is in the form A – B – C, where C is always 96.

Identifying the Number Series Pattern

Since 96 is constant, the pattern likely involves the relationship between the first two numbers (A and B) and possibly 96. Let's explore a potential pattern related to the difference between A and B and how it relates to 96.

Consider the absolute difference between the first number (A) and the second number (B), which is $|A - B|$. Let's check if this difference has a specific relationship with 96, such as being a factor of 96.

First, let's list the factors of 96. Factors are numbers that divide 96 evenly.

$$\text{Factors of 96} : \{1, 2, 3, 4, 6, 8, 12, 16, 24, 32, 48, 96\}$$

Now, let's calculate the absolute difference $|A - B|$ for each element in the series and check if it is a factor of 96.

Checking Each Element for the Pattern

Element	First No. (A)	Second No. (B)	Absolute Difference $\|A - B\|$	Is $\|A - B\|$ a Factor of 96?	Fits Pattern?
5 – 1 – 96	5	1	$\|5 - 1\| = 4$	Yes (4 is in the factors list)	Yes
27 – 1 – 96	27	1	$\|27 - 1\| = 26$	No (26 is not in the factors list)	No
18 – 2 – 96	18	2	$\|18 - 2\| = 16$	Yes (16 is in the factors list)	Yes
12 – 3 – 96	12	3	$\|12 - 3\| = 9$	No (9 is not in the factors list)	No
2 – 4 – 96	2	4	$\|2 - 4\| = 2$	Yes (2 is in the factors list)	Yes

Pinpointing the Outlier in the Series

Based on the pattern that the absolute difference between the first two numbers ($|A - B|$) should be a factor of 96, most elements in the series fit this rule:

- 5 – 1 – 96 fits because $|5 - 1| = 4$, and 4 is a factor of 96.
- 18 – 2 – 96 fits because $|18 - 2| = 16$, and 16 is a factor of 96.
- 2 – 4 – 96 fits because $|2 - 4| = 2$, and 2 is a factor of 96.

However, two elements do not fit this pattern:

- 27 – 1 – 96 does not fit because $|27 - 1| = 26$, and 26 is not a factor of 96.
- 12 – 3 – 96 does not fit because $|12 - 3| = 9$, and 9 is not a factor of 96.

The question asks which of the given options doesn't fit the series. Among the provided options, 12 – 3 – 96 is listed. Our analysis shows that 12 – 3 – 96 does not follow the pattern where the absolute difference between the first two numbers is a factor of 96.

Conclusion: Finding the Outlier

The pattern observed in the series is that the absolute difference between the first and second number is a factor of 96. The element 12 – 3 – 96 has an absolute difference of $|12 - 3| = 9$, and 9 is not a factor of 96. Therefore, this element doesn't fit the established pattern of the series.

Number Series Revision Table

Review of absolute differences and factors of 96:

Element	$\lvert A - B \rvert$	Is it a factor of 96?
5 – 1 – 96	4	Yes
27 – 1 – 96	26	No
18 – 2 – 96	16	Yes
12 – 3 – 96	9	No
2 – 4 – 96	2	Yes

Additional Information on Factors and Divisibility

A factor of a number is an integer that divides the number without leaving a remainder. For example, the factors of 12 are 1, 2, 3, 4, 6, and 12, because 12 can be divided evenly by these numbers ($12 \div 1 = 12$, $12 \div 2 = 6$, etc.).

Divisibility is the property of a number being divisible by another number without a remainder. We checked if 96 is divisible by the absolute difference $\lvert A - B \rvert$. If $96 \div \lvert A - B \rvert$ results in a whole number (integer) with no remainder, then $\lvert A - B \rvert$ is a factor of 96.

For instance, for 5 – 1 – 96, $\lvert A - B \rvert = 4$. $96 \div 4 = 24$. Since 24 is a whole number, 4 is a factor of 96.

For 12 – 3 – 96, $\lvert A - B \rvert = 9$. $96 \div 9 = 10.66\ldots$. Since this is not a whole number, 9 is not a factor of 96.

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16. **Answer: d**

Explanation:

Understanding Analogies: Teacher, Class, Driver

This question is an example of a verbal analogy. Analogies test your ability to find the relationship between a pair of words and then apply that same relationship to another word to find the missing term. In the analogy "Teacher is related to Class in the same way Driver is related to _____," we need to determine the connection between a Teacher and a Class and then find the word that shares the identical relationship with a Driver.

Identifying the Relationship: Teacher and Class

Let's examine the relationship between the first pair: **Teacher** and **Class**.

- A Teacher works or teaches **in** a Class.
- A Class is the place where a Teacher performs their primary job function.
- The relationship is essentially "Person : Place of Work/Activity".

Applying the Relationship: Driver and the Missing Term

Now we apply the same "Person : Place of Work/Activity" relationship to the second pair, starting with **Driver**. A Driver is a person who performs the action of driving. Where does a Driver perform their main activity?

Analyzing the Options for the Driver Analogy

Let's look at the given options:

- **Brake:** A brake is a part of a vehicle, a tool used by the driver. It's not the place where the driver performs their activity.
- **Wheel:** A wheel is also a part of a vehicle. Similar to the brake, it's a component the driver interacts with, but not the location of the driving activity itself.
- **Parts:** This is a general term referring to components of something, likely a vehicle. It's too general and doesn't represent the place or entity the driver operates.
- **Vehicle:** A vehicle (like a car, bus, truck, etc.) is the entity that a driver operates or drives. The driver performs their job **in** or **on** a Vehicle. This aligns perfectly with the relationship we found: "Person : Place of Work/Activity" (Driver : Vehicle).

Conclusion: Solving the Analogy

The relationship between Teacher and Class is that a Teacher works in a Class. Applying this relationship, a Driver works in or on a Vehicle. Therefore, Driver is related to Vehicle in the same way Teacher is related to Class.

The correct option that completes the analogy is **Vehicle**.

Revision Table: Analogy Breakdown

First Pair (Known)	Relationship	Second Pair (Analogy)
Teacher : Class	Person works in/at Place	Driver : ?
Teacher works in Class	Same relationship applied	Driver works in/on Vehicle

Additional Information: Types of Analogies

Analogies can be based on various relationships. Some common types include:

- **Part to Whole:** Finger : Hand (Finger is part of a Hand)
- **Cause and Effect:** Study : Success (Studying can cause Success)
- **Opposites:** Hot : Cold (Hot is the opposite of Cold)
- **Synonyms:** Happy : Joyful (Happy and Joyful are synonyms)
- **Performer and Action:** Chef : Cook (A Chef performs the action of Cooking)
- **Performer and Object of Action:** Hunter : Prey (A Hunter hunts Prey)
- **Performer and Location:** Doctor : Hospital (A Doctor works in a Hospital) - similar to the question's analogy
- **Tool and User:** Pen : Writer (A Pen is used by a Writer)

Understanding these different types helps in solving various analogy questions in verbal reasoning tests.

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17. **Answer: a**

Explanation:

Solving Mathematical Expressions Using Order of Operations

To solve the given mathematical expression, we need to follow the order of operations, often remembered by the acronym BODMAS or PEMDAS. This order ensures that we perform calculations in the correct sequence to arrive at the accurate result.

The expression is:

$(55 \div 11) + (18–6) \times 9$

Understanding BODMAS/PEMDAS

BODMAS stands for:

- **B**rackets first
- **O**rders (powers, square roots, etc.)
- **D**ivision and **M**ultiplication (from left to right)
- **A**ddition and **S**ubtraction (from left to right)

PEMDAS stands for Parentheses, Exponents, Multiplication and Division, Addition and Subtraction.

Let's apply this order to our expression.

Step-by-Step Calculation

Step 1: Solve the operations inside the Brackets.

There are two sets of brackets in the expression:

$(55 \div 11)$

$(18–6)$

Calculate the first bracket:

$55 \div 11 = 5$

Calculate the second bracket:

$18–6 = 12$

Now, substitute these results back into the original expression:

$5 + 12 \times 9$

Step 2: Perform Multiplication.

According to BODMAS/PEMDAS, Multiplication comes before Addition. We have 12×9.

$12 \times 9 = 108$

Substitute this result back into the expression:

$5 + 108$

Step 3: Perform Addition.

Finally, perform the addition:

$5 + 108 = 113$

So, the value of the expression $(55 \div 11) + (18–6) \times 9$ is 113.

Summary of Operations

Operation	Calculation	Expression
Original Expression		$(55 \div 11) + (18–6) \times 9$
Brackets $(55 \div 11)$	$55 \div 11 = 5$	$5 + (18–6) \times 9$
Brackets $(18–6)$	$18–6 = 12$	$5 + 12 \times 9$
Multiplication	$12 \times 9 = 108$	$5 + 108$
Addition	$5 + 108 = 113$	113

The final answer is 113.

Revision Table: Order of Operations

Order	Operation Type	Description
1	Brackets / Parentheses	Perform calculations inside ( ), { }, [ ]
2	Orders / Exponents	Calculate powers, square roots, etc.
3	Division and Multiplication	Perform from left to right
4	Addition and Subtraction	Perform from left to right

Additional Information on Mathematical Expressions

Mathematical expressions are combinations of numbers, variables, operators (like +, −, ×, ÷), and grouping symbols (like brackets). Evaluating an expression means finding its numerical value.

- It's crucial to follow the order of operations consistently to avoid errors.
- If an expression has multiple divisions or multiplications, perform them in the order they appear from left to right. The same rule applies to multiple additions or subtractions.
- Parentheses or brackets can change the usual order of operations by forcing certain calculations to be done first.

Understanding the order of operations is fundamental for solving more complex mathematical problems in algebra and beyond.

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18. **Answer: b**

Explanation:

Understanding Weight on the Moon vs. Earth

The question asks about the relationship between the weight of an object on the Moon compared to its weight on Earth. To understand this, we need to consider what weight is and how gravity affects it.

What is Weight?

Weight is the force of gravity acting on an object's mass. It is calculated using the formula:

$W = m \times g$

Where:

- W is the weight of the object
- m is the mass of the object
- g is the acceleration due to gravity

Mass (m) is a measure of the amount of matter in an object, and it remains the same regardless of location (whether on Earth, the Moon, or in space). However, the acceleration due to gravity (g) varies depending on the celestial body's mass and size.

Gravity on Earth and the Moon

The Earth is much more massive than the Moon. This larger mass creates a stronger gravitational pull. The approximate value for the acceleration due to gravity on Earth (g_{earth}) is $9.8\,\mathrm{m/s^2}$.

The Moon has significantly less mass than Earth. Consequently, its gravitational pull is weaker. The approximate value for the acceleration due to gravity on the Moon (g_{moon}) is about $1.62\,\mathrm{m/s^2}$.

Comparing Gravitational Acceleration

Let's compare the gravitational acceleration on the Moon to that on Earth:

$\frac{g_{moon}}{g_{earth}} \approx \frac{1.62\,\mathrm{m/s^2}}{9.8\,\mathrm{m/s^2}} \approx \frac{1}{6}$

This shows that the gravitational acceleration on the Moon is approximately one-sixth $(1/6)$ of the gravitational acceleration on Earth.

Relating Weight on Moon and Earth

Since weight is directly proportional to the acceleration due to gravity $(W = m \times g)$, and the mass (m) of the object stays constant, the ratio of the weight on the Moon to the weight on Earth is the same as the ratio of their gravitational accelerations.

Weight on the Moon: $W_{moon} = m \times g_{moon}$

Weight on Earth: $W_{earth} = m \times g_{earth}$

Ratio of weights:

$$\frac{W_{moon}}{W_{earth}} = \frac{m \times g_{moon}}{m \times g_{earth}} = \frac{g_{moon}}{g_{earth}}$$

Since $\frac{g_{moon}}{g_{earth}} \approx \frac{1}{6}$, it follows that:

$$\frac{W_{moon}}{W_{earth}} \approx \frac{1}{6}$$

Or, $W_{moon} \approx \frac{1}{6} \times W_{earth}$

Therefore, the weight of an object on the moon is approximately one-sixth $(1/6)$ the weight of the object on Earth.

Location	Gravitational Acceleration (g)	Weight (for a given mass m)
Earth	$\approx 9.8\,\mathrm{m/s}^2$	$W_{earth} = m \times g_{earth}$
Moon	$\approx 1.62\,\mathrm{m/s}^2$	$W_{moon} = m \times g_{moon}$

Comparing the options:

- equal to: This is incorrect. Gravity is different.
- 1/6th: This aligns with our calculation based on the ratio of gravitational acceleration.
- 1/2: This is incorrect. The ratio is closer to 1/6.

- 1/5th: This is incorrect. The ratio is closer to 1/6.

Based on the principles of gravity and weight, the weight of an object on the moon is 1/6th the weight of the object on earth.

Revision Table: Weight on Moon vs Earth

Concept	Earth	Moon	Comparison (Moon vs Earth)
Mass	m	m	Same
Acceleration due to Gravity (g)	$\approx 9.8\,\mathrm{m/s^2}$	$\approx 1.62\,\mathrm{m/s^2}$	$g_{moon} \approx 1/6 \times g_{earth}$
Weight ($W = m \times g$)	W_{earth}	W_{moon}	$W_{moon} \approx 1/6 \times W_{earth}$

Additional Information: Mass and Weight

It is important to distinguish between mass and weight:

- **Mass:** This is the amount of matter in an object. It is an intrinsic property of the object and does not change with location. Mass is typically measured in kilograms (kg).
- **Weight:** This is the force exerted on an object due to gravity. It depends on both the object's mass and the strength of the gravitational field it is in. Weight is a force and is measured in Newtons (N). On Earth, mass is often colloquially used interchangeably with weight (e.g., "I weigh 70 kg"), but scientifically, kilograms measure mass, and weight is a force.

Because the Moon has less mass and a smaller radius compared to Earth, its gravitational pull is significantly weaker. This weaker gravity is the direct reason why objects weigh less on the Moon.

Astronauts on the Moon experience this reduced gravity directly, finding it much easier to move and even leap large distances compared to on Earth, despite their mass remaining unchanged.

19. **Answer: a**

Explanation:

Understanding the Work and Time Problem

This problem involves the concept of work, which is often considered proportional to the number of workers and the time they spend. A common way to represent total work is by the product of the number of people and the number of days (or hours, etc.) they work. We can call this unit 'person-days' or 'worker-days'.

Initial Work Calculation

We are told that 48 people could complete the entire task in 17 days. Assuming the rate of work per person is constant, the total amount of work required for the task is:

$$\text{Total Work} = \text{Number of People} \times \text{Time (Days)}$$

$$\text{Total Work} = 48 \text{ people} \times 17 \text{ days}$$

$$\text{Total Work} = 816 \text{ worker-days}$$

Work Done in the First 6 Days

The 48 people worked for the first 6 days. The amount of work completed during this period is:

$$\text{Work Done} = \text{Number of People} \times \text{Time (Days)}$$

$$\text{Work Done} = 48 \text{ people} \times 6 \text{ days}$$

$$\text{Work Done} = 288 \text{ worker-days}$$

Calculating Remaining Work

After 6 days, a portion of the task is complete. The remaining work is the total work minus the work already done:

$$\text{Remaining Work} = \text{Total Work} - \text{Work Done}$$

$$\text{Remaining Work} = 816 \text{ worker-days} - 288 \text{ worker-days}$$

$$\text{Remaining Work} = 528 \text{ worker-days}$$

Calculating Remaining Workers

After 6 days, 4 of the workers left. The number of workers remaining to complete the task is:

$$\text{Remaining Workers} = \text{Initial Workers} - \text{Workers Who Left}$$

$$\text{Remaining Workers} = 48 - 4$$

$$\text{Remaining Workers} = 44 \text{ people}$$

Time to Complete Remaining Work

Now, 44 workers need to complete the remaining 528 worker-days of work. To find out how many days this will take, we divide the remaining work by the number of remaining workers:

$$\text{Days to Complete Remaining Work} = \frac{\text{Remaining Work}}{\text{Remaining Workers}}$$

$$\text{Days to Complete Remaining Work} = \frac{528 \text{ worker-days}}{44 \text{ people}}$$

$$\text{Days to Complete Remaining Work} = 12 \text{ days}$$

So, it will take 12 days from then for the remaining 44 workers to complete the task.

Description	Value	Calculation/Reason
Initial Workers	48	Given
Initial Total Days	17	Given
Total Work (worker-days)	816	48×17
Days Worked Initially	6	Given
Work Done in 6 Days (worker-days)	288	48×6
Workers Who Left	4	Given
Remaining Workers	44	$48 - 4$
Remaining Work (worker-days)	528	$816 - 288$
Days to Complete Remaining Work	12	$528/44$

Therefore, the remaining task will be completed in 12 days from the time the 4 workers left.

Revision Table: Work and Time Concepts

Concept	Formula/Explanation
Total Work	Often calculated as Manpower $\times$ Time (e.g., worker-days)
Work Rate	Work done per unit of time by a person or a group. Assumed constant per person unless stated otherwise.
Relation between Men, Work, and Time	If M workers can do W work in T time, then $\frac{M_1 T_1}{W_1} = \frac{M_2 T_2}{W_2}$. If work is the same $(W_1 = W_2)$, then $M_1 T_1 = M_2 T_2$. This is an inverse proportion: more workers mean less time for the same work.
Partial Work	Work done in a specific period by a specific number of workers.
Remaining Work	Total Work - Work Done.

Additional Information on Time and Work Problems

Problems involving time and work often assume that all workers work at the same rate unless specified otherwise. The core idea is that the total amount of work required for a task is fixed. If the number of workers changes, the time taken to complete the same amount of work will change inversely.

- If more people work, the time taken decreases.
- If fewer people work, the time taken increases.
- The unit of work (like worker-days) helps quantify the total effort needed.
- Calculating the work done and remaining work is key when there are changes in the number of workers or their efficiency during the task.

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20. **Answer: b**

Explanation:

Understanding Subhash Chandra Bose's Political Party in 1939

The question asks about the political party founded by Subhash Chandra Bose in the year 1939. Subhash Chandra Bose was a prominent leader in the Indian independence movement. After differences arose within the Indian National Congress, particularly following his re-election as president in 1939 and subsequent resignation, he decided to form his own political group.

Formation of the Forward Bloc

In 1939, after resigning from the presidency of the Indian National Congress, Subhash Chandra Bose launched a new political entity. This party aimed to consolidate the left-wing elements within the Congress and work towards immediate independence. The name of this party is the **All India Forward Bloc**.

Let's look at the options provided:

- Communist Party of India: This party was formed much earlier and had a different origin and leadership.
- All India Forward Bloc: This party was indeed founded by Subhash Chandra Bose in 1939.
- Azad Bengal Fauj: This name sounds similar to military organizations associated with Bose, like the Azad Hind Fauj (Indian National Army), but it's not the name of the political party he launched in 1939.
- Socialist Party of India: Various socialist groups existed, but the specific party launched by Bose in 1939 was the Forward Bloc.

Based on historical facts, the party launched by Subhash Chandra Bose in 1939 was the All India Forward Bloc.

Key Details: Subhash Chandra Bose and Forward Bloc

Leader	Party Launched	Year	Context
Subhash Chandra Bose	All India Forward Bloc	1939	After resigning from Indian National Congress presidency

Conclusion on Subhash Chandra Bose's Party

The party established by Subhash Chandra Bose after leaving the Indian National Congress in 1939 was the All India Forward Bloc. This organization became a significant part of the Indian political landscape, continuing its activities even after Bose's eventual departure from India.

Revision Table: Important Parties

Indian Political Parties and Founders/Associated Leaders

Party	Key Figure(s)	Formation Year (Approx.)
Indian National Congress	A.O. Hume, various leaders	1885
Muslim League	Various leaders	1906
Communist Party of India	M.N. Roy (early influence)	1920s
All India Forward Bloc	Subhash Chandra Bose	1939

Additional Information: Subhash Chandra Bose

Subhash Chandra Bose was a dynamic leader who advocated for complete and immediate independence for India. He served as the President of the Indian National Congress in 1938 (Haripura Session) and was re-elected in 1939 (Tripuri Session). However, due to ideological differences with Mahatma Gandhi and the Congress Working Committee, he resigned from the presidency.

Following his resignation, he formed the Forward Bloc within the Congress in May 1939. The Bloc aimed to rally radical elements and continue the struggle for immediate independence. Later, the Forward Bloc was removed from the Congress. Bose's journey after 1939 involved seeking international support for India's independence, famously leading to the formation of the Provisional Government of Free India and the Indian National Army (Azad Hind Fauj).

Understanding the formation of the All India Forward Bloc in 1939 is key to tracing the political journey of Subhash Chandra Bose during this critical period of the independence movement.

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21. **Answer: b**

Explanation:

Understanding Plant Tissues and Their Functions

Plants have different types of tissues that perform specific jobs. Some tissues help with transport, others with photosynthesis, and some provide support and structure. The question asks which tissue makes the plant hard and stiff.

Analysis of Plant Tissue Options

Let's look at the different options provided:

- **Parenchyma:** This is a basic plant tissue. Parenchyma cells are living and usually have thin cell walls. They are involved in storage, photosynthesis, and secretion. While they provide some support when turgid (full of water), they do not make the plant hard and stiff in a permanent way.
- **Sclerenchyma:** This tissue is specifically designed for mechanical support and strength. Sclerenchyma cells have thick, lignified (hardened by lignin) secondary cell walls. These cells are often dead at maturity. The thick, rigid walls provide significant stiffness and make parts of the plant very hard. Examples include the hard shell of nuts, the gritty texture of pears, and the fibers in hemp or jute.
- **Collenchyma:** This tissue provides mechanical support to growing parts of the plant, like young stems and petioles. Collenchyma cells have unevenly thickened primary cell walls, usually at the corners. They are living cells and are flexible, allowing the plant to bend without breaking. While they offer support, they don't provide the hardness and stiffness that sclerenchyma does.
- **Xylem:** This is a complex vascular tissue responsible for transporting water and dissolved minerals from the roots to the rest of the plant. Xylem also provides some mechanical support due to the lignified walls of some of its components (like vessels and tracheids). However, its primary role is transport, not overall hardness and stiffness of structures like fruit shells or fibers.

Sclerenchyma: The Tissue for Hardness and Stiffness

Based on the functions of these tissues, sclerenchyma is the tissue that makes the plant hard and stiff. Its characteristic feature is the presence of heavily thickened and often lignified cell walls. Lignin is a complex polymer that provides rigidity and strength to the plant cell wall.

Sclerenchyma tissue is found in various parts of the plant where strength and rigidity are needed, such as:

- In the stems and leaves to provide structural support.
- Around vascular bundles.
- In the hard coverings of seeds and nuts.
- In the gritty flesh of some fruits.

Sclerenchyma exists in two main forms:

- **Fibers:** Long, slender cells, often found in bundles, providing strength to stems and leaves (e.g., flax, hemp).
- **Sclereids:** Variously shaped cells (often shorter than fibers), providing hardness to seed coats, nut shells, and fruit pulp (e.g., stone cells in pears).

The rigidity provided by the thick, lignified walls of sclerenchyma cells is what makes plant parts hard and stiff.

Comparison of Plant Tissues and Function

Tissue Type	Cell Type	Cell Wall	Main Function
Parenchyma	Living	Thin primary wall	Storage, Photosynthesis, Secretion
Collenchyma	Living	Unevenly thickened primary wall	Support for growing parts, Flexibility
Sclerenchyma	Usually Dead at Maturity	Thick, lignified secondary wall	Mechanical Support, Hardness, Stiffness
Xylem	Mix (some dead)	Lignified walls (in vessels/tracheids)	Water & mineral transport, Some support

Therefore, the tissue primarily responsible for making the plant hard and stiff is sclerenchyma.

Revision Table: Key Plant Tissues

Summary of Plant Tissues

Tissue	Structure	Role in Support/Structure
Parenchyma	Thin-walled living cells	Turgor pressure support (temporary)
Collenchyma	Unevenly thickened living cells	Support for growing stems/leaves (flexible)
Sclerenchyma	Thick, lignified, often dead cells (fibers, sclereids)	Provides rigidity, hardness, stiffness, mechanical strength
Xylem	Complex tissue (vessels, tracheids, etc.)	Water transport; Lignification adds some structural support

Additional Information on Structural Plant Tissues

Plant tissues are broadly classified into simple tissues (made of one type of cell) and complex tissues (made of more than one type of cell). Parenchyma, collenchyma, and sclerenchyma are simple tissues. Xylem and phloem are complex tissues.

Sclerenchyma provides mechanical support and is crucial for the survival of the plant in challenging environments, helping it withstand forces like wind. The strength provided by sclerenchyma fibers has also made them economically important for humans, used in textiles, ropes, and other materials for centuries.

The cell walls of sclerenchyma are much thicker than those of parenchyma or collenchyma. This thickness and the presence of lignin are key to their function in providing hardness and stiffness.

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22. **Answer: b**

Explanation:

Finding Students Taking Both Subjects: A Set Theory Approach

This problem involves calculating the number of students who took both Mathematics and Biology using the principles of set theory. We are given the total number of students, the number of students in each subject group, and the number of students who took neither subject.

Understanding the Given Information

Let's list the information provided:

- Total number of students in the class: 60
- Number of students who took Mathematics: $|M| = 29$
- Number of students who took Biology: $|B| = 32$
- Number of students who took none of the two subjects: 8

We need to find the number of students who took both Mathematics and Biology, which corresponds to the intersection of the two sets, $|M \cap B|$.

Calculating Students Taking At Least One Subject

First, we can find the number of students who took at least one of the two subjects (Mathematics or Biology). This includes students who took only Mathematics, only Biology, or both. We can find this by subtracting the number of students who took none from the total number of students.

Number of students taking at least one subject = Total students - Students taking none

Number of students taking at least one subject $= 60 - 8 = 52$.

In set theory terms, this is the union of the two sets, $|M \cup B|$. So, $|M \cup B| = 52$.

Applying the Principle of Inclusion-Exclusion

For two sets, the principle of inclusion-exclusion states that the number of elements in the union of two sets is equal to the sum of the number of elements in each set minus the number of elements in their intersection.

The formula is:

$$|M \cup B| = |M| + |B| - |M \cap B|$$

We know the values for $|M \cup B|$, $|M|$, and $|B|$. We can plug these values into the formula to solve for $|M \cap B|$, the number of students who took both subjects.

$$52 = 29 + 32 - |M \cap B|$$

Now, let's perform the calculation:

First, sum the number of students in Mathematics and Biology:

$$29 + 32 = 61$$

So the equation becomes:

$$52 = 61 - |M \cap B|$$

To find $|M \cap B|$, we rearrange the equation:

$$|M \cap B| = 61 - 52$$

$$|M \cap B| = 9$$

Therefore, 9 students took both Mathematics and Biology.

Summary of Steps

Step	Description	Calculation
1	Find students taking at least one subject ($|M \cup B|$).	$60 - 8 = 52$
2	Use Inclusion-Exclusion principle: $|M \cup B| = |M| + |B| - |M \cap B|$.	$52 = 29 + 32 - |M \cap B|$
3	Solve for $|M \cap B|$.	$|M \cap B| = 29 + 32 - 52 = 61 - 52 = 9$

The number of students who took both Mathematics and Biology is 9.

Revision Table: Key Concepts Review

Term	Definition/Concept	Relation to Problem
Set Union ($A \cup B$)	Elements belonging to set A OR set B (or both). Represents students taking Mathematics OR Biology (or both).	Calculated as Total - None = 52.
Set Intersection ($A \cap B$)	Elements belonging to set A AND set B. Represents students taking BOTH Mathematics AND Biology.	This is what we needed to find ($\lvert M \cap B\rvert$).
Principle of Inclusion-Exclusion	$\lvert A \cup B\rvert = \lvert A\rvert + \lvert B\rvert - \lvert A \cap B\rvert$ for two sets. Helps relate union, individual sets, and intersection.	Used directly to solve for $\lvert M \cap B\rvert$.

Additional Information: Venn Diagrams

Problems like this can also be visualized using a Venn diagram. A Venn diagram for two sets, M (Mathematics) and B (Biology), would have two overlapping circles inside a rectangle representing the total class.

- The overlapping region represents $|M \cap B|$ (both subjects).
- The part of circle M only represents students taking only Mathematics ($|M| - |M \cap B|$).
- The part of circle B only represents students taking only Biology ($|B| - |M \cap B|$).
- The area outside the circles but inside the rectangle represents students taking neither subject.

In this case, the sum of all these regions must equal the total number of students:

$$(\text{Only M}) + (\text{Only B}) + (\text{Both}) + (\text{None}) = \text{Total}$$

$$(|M| - |M \cap B|) + (|B| - |M \cap B|) + |M \cap B| + (\text{None}) = \text{Total}$$

This simplifies to:

$$|M| + |B| - |M \cap B| + (\text{None}) = \text{Total}$$

Rearranging to solve for $|M \cap B|$:

$$|M \cap B| = |M| + |B| + (\text{None}) - \text{Total}$$

Let's check this using the numbers:

$$|M \cap B| = 29 + 32 + 8 - 60$$

$$|M \cap B| = 61 + 8 - 60$$

$$|M \cap B| = 69 - 60$$

$$|M \cap B| = 9$$

Both the inclusion-exclusion formula and the Venn diagram approach yield the same result, confirming the answer that 9 students took both subjects.

---

23. **Answer: a**

Explanation:

Understanding Metal Properties: Malleability for Metallic Foils

The question asks about the specific property of metals that allows them to be made into thin sheets, commonly known as metallic foils. Let's examine the given options to understand which property is responsible for this.

Metals possess several physical properties that make them useful for various applications. The ability to transform a metal into a particular shape or form depends heavily on these inherent properties.

Analyzing the Properties of Metals

- **Malleability:** This is the property of a material that allows it to be hammered, pressed, or rolled into thin sheets without breaking. Metals like gold, silver, aluminum, and copper are known for their high malleability. Making metallic foils, such as aluminum foil used in packaging, is a direct application of this property.

- **Ductility:** This is the property of a material that allows it to be drawn or stretched into thin wires. While some metals are both malleable and ductile (like gold), ductility specifically relates to forming wires, not thin sheets.
- **Sonority:** This refers to the property of a material that causes it to produce a ringing sound when struck. Metals are generally sonorous, which is why bells are made of metal, but this property is unrelated to shaping the metal into foils.
- **Conductivity:** This refers to the ability of a material to conduct heat (thermal conductivity) or electricity (electrical conductivity). Metals are typically good conductors, which makes them useful in electrical wiring and cookware, but this property does not explain how foils are formed.

Connecting the Property to Metallic Foils

Metallic foils are essentially very thin sheets of metal. The process of creating these thin sheets involves significant deformation of the metal, often by rolling or hammering. The property that allows a material to undergo such plastic deformation and be flattened into a sheet without fracturing is malleability.

Therefore, the production of metallic foils is directly dependent on the malleability of the metal.

Comparison of Properties

Property	Definition	Application Example
Malleability	Can be hammered/rolled into thin sheets	Metallic foils (e.g., aluminum foil), metal coins
Ductility	Can be drawn into wires	Electrical wires (e.g., copper wire)
Sonority	Produces a ringing sound when struck	Bells
Conductivity	Allows heat or electricity to pass through	Electrical wiring, cooking utensils

Based on the definitions and examples, it is clear that malleability is the property that enables the production of metallic foils.

Conclusion

The property of metals that enables the production of metallic foils is malleability. This allows metals to be shaped into thin sheets.

Revision Table: Metal Properties and Uses

Metal Property	Enables Production Of	Examples
Malleability	Thin sheets (foils)	Aluminum foil, gold leaf
Ductility	Thin wires	Copper wires, gold threads
Sonority	Sound production	Bells, musical instruments
Conductivity (Thermal/Electrical)	Heat/Electricity transfer	Cookware, electrical circuits

Additional Information on Malleability

Some metals are more malleable than others. Gold is one of the most malleable metals; it can be hammered into extremely thin sheets called gold leaf, often used in decoration. Aluminum is also highly malleable, which is why aluminum foil is commonly used in kitchens and packaging. Copper and silver are also quite malleable. Iron is malleable when hot, allowing it to be forged into various shapes, but it is less malleable than gold or aluminum at room temperature.

Malleability is an important physical property for many industrial processes involving shaping metals without melting them, such as rolling, stamping, and forging, all of which involve deforming the metal under pressure.

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24. **Answer: d**

Explanation:

Understanding the Letter-Based Mathematical Expression

This question presents a mathematical expression encoded using letters instead of standard arithmetic symbols. To find the value of the expression, we first need to decode the letters back into their corresponding mathematical operations and then evaluate the resulting expression following the correct order of operations.

Decoding the Symbols for Evaluation

The question provides the following mapping for the letters:

- 'Q' means '+' (Addition)
- 'J' means '×' (Multiplication)
- 'T' means '−' (Subtraction)
- 'K' means '÷' (Division)

Translating the Given Expression

The expression provided is:

$18K3Q7J2T8$

Using the decoding information, we substitute the letters with the standard mathematical operators:

- 'K' is replaced by '÷'
- 'Q' is replaced by '+'
- 'J' is replaced by '×'
- 'T' is replaced by '–'

The expression becomes:

$18 \div 3 + 7 \times 2 - 8$

Evaluating the Mathematical Expression Using Order of Operations

To calculate the value of the translated expression $18 \div 3 + 7 \times 2 - 8$, we must follow the standard mathematical order of operations, often remembered as BODMAS or PEMDAS.

The order is generally: Brackets/Parentheses, Orders/Exponents, Division and Multiplication (from left to right), and Addition and Subtraction (from left to right).

Let's apply this order step-by-step:

Step 1: Perform Division and Multiplication

According to the order of operations, division and multiplication are performed before addition and subtraction. We work from left to right.

First, the division:

$18 \div 3 = 6$

The expression is now:

$6 + 7 \times 2 - 8$

Next, the multiplication:

$7 \times 2 = 14$

The expression is now:

$6 + 14 - 8$

Step 2: Perform Addition and Subtraction

Now, we perform addition and subtraction from left to right.

First, the addition:

$6 + 14 = 20$

The expression is now:

$20 - 8$

Finally, the subtraction:

$20 - 8 = 12$

The value of the expression $18 \div 3 + 7 \times 2 - 8$ is 12.

Conclusion of Expression Evaluation

By correctly decoding the letter-based expression and applying the standard order of operations, we found that the value of $18K3Q7J2T8$ is 12.

Revision Table: Summary of Symbol Mapping

Letter	Operation	Symbol
Q	Addition	+
J	Multiplication	×
T	Subtraction	-
K	Division	÷

Additional Information: Order of Operations (BODMAS/PEMDAS)

The order of operations is a fundamental concept in arithmetic and algebra. It dictates the sequence in which mathematical operations should be performed to evaluate an expression unambiguously. Without this standard order, the outcome of a calculation could vary depending on the interpreter. The common acronyms BODMAS and PEMDAS help remember this order:

- **BODMAS:** Brackets, Orders (powers/roots), Division and Multiplication, Addition and Subtraction.
- **PEMDAS:** Parentheses, Exponents, Multiplication and Division, Addition and Subtraction.

It is important to remember that Division and Multiplication are performed from left to right, and similarly, Addition and Subtraction are performed from left to right. They do not have strict precedence over each other, only over operations lower on the list.

Mastering the order of operations is essential for solving various mathematical problems accurately, including those involving symbol substitution as seen in this question.

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25. Answer: c

Explanation:

Understanding Deliquescence and Identifying Deliquescent Compounds

The question asks to identify a deliquescent compound from the given options. To answer this, let's first understand what a deliquescent compound is.

What is a Deliquescent Compound?

A deliquescent compound is a substance, typically a salt, that is so highly hygroscopic (meaning it readily absorbs moisture from the atmosphere) that it absorbs enough water to dissolve in the absorbed water and form a solution. This process requires the vapor pressure of the saturated solution formed to be lower than the partial pressure of water vapor in the surrounding air.

Let's examine the properties of each substance provided in the options in relation to moisture absorption:

- **Quick lime (Calcium oxide, CaO):** Quick lime is hygroscopic and readily absorbs moisture from the air. However, it primarily reacts with water to form calcium hydroxide ($Ca(OH)_2$), a process called slaking. While it absorbs moisture, it doesn't typically dissolve into a solution under normal conditions in the same way that typical deliquescent salts do.
- **Sulphuric acid (H_2SO_4):** Concentrated sulphuric acid is a powerful dehydrating agent, meaning it has a very strong affinity for water and absorbs water vapor from the air. However, it is already a liquid, and the term deliquescence usually refers to a solid absorbing moisture and turning into a liquid solution. So, while it absorbs moisture strongly, it's not classified as a deliquescent compound in the same sense as a solid salt.
- **Phosphorous pentoxide (P_4O_{10} or P_2O_5):** Phosphorous pentoxide is an extremely powerful drying agent. It reacts vigorously with water to form phosphoric acid (H_3PO_4). It absorbs moisture from the air so effectively that it often liquefies or turns into a syrupy liquid due to the reaction with absorbed water, making it behave like a deliquescent substance in terms of strong

moisture absorption leading to a liquid state. It is commonly cited as a substance exhibiting deliquescent properties due to its intense hygroscopy.
- **Sodium chloride (Salt, $NaCl$):** Sodium chloride is slightly hygroscopic, meaning it can absorb small amounts of moisture, especially in very humid conditions, which is why salt might clump together. However, it is not considered deliquescent under normal atmospheric conditions because it does not absorb enough moisture to dissolve completely into a solution unless the humidity is extremely high.

Comparing these properties, Phosphorous pentoxide is known for its extremely high affinity for moisture, leading to its reaction and effective liquefaction upon exposure to air. This strong moisture absorption causing it to become liquid fits the description of a substance exhibiting deliquescent behaviour, even if it involves a chemical reaction rather than just simple dissolution.

Summary of Properties

Substance	Nature regarding moisture	Deliquescent?
Quick lime (CaO)	Hygroscopic (reacts with water)	No (Reacts, doesn't dissolve to solution)
Sulphuric acid (H_2SO_4)	Powerful dehydrating agent (absorbs water)	No (Already liquid, absorbs water)
Phosphorous pentoxide (P_4O_{10})	Extremely hygroscopic (reacts vigorously with water, becomes liquid)	Yes (Behaves like deliquescent due to extreme water absorption)
Sodium chloride ($NaCl$)	Slightly hygroscopic	No (Not under normal conditions)

Based on the properties, Phosphorous pentoxide is the compound that is considered deliquescent or behaves in a deliquescent manner due to its strong ability to absorb moisture from the air and become a liquid.

Revision Table: Key Concepts

Term	Definition
Hygroscopic	Property of a substance to absorb moisture from the air.
Deliquescence	Extreme hygroscopy where a substance absorbs enough moisture to dissolve in it and form a solution.
Efflorescence	Property of a substance to lose water of crystallization when exposed to air.

Additional Information: Examples of Deliquescent Compounds

Besides Phosphorous pentoxide, other common examples of deliquescent compounds include:

- Calcium chloride $(CaCl_2)$
- Sodium hydroxide $(NaOH)$
- Potassium hydroxide (KOH)
- Ferric chloride $(FeCl_3)$
- Magnesium chloride $(MgCl_2)$

These substances are often used as drying agents because of their ability to remove moisture from the air or other substances.

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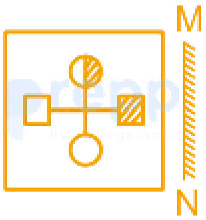
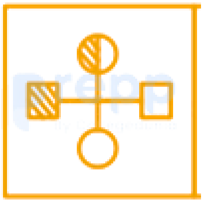
26. **Answer: c**

Explanation:





Hence, the figure A is the correct mirror image.

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27. **Answer: b**

Explanation:

Understanding the Establishment of the Bombay Stock Exchange (BSE)

The question asks about the year in which the Bombay Stock Exchange (BSE) was established. The BSE is one of the oldest and most significant stock exchanges in India and Asia.

Let's look at the options provided:

- 1947: This year is historically important for India (Independence) but not for the BSE's establishment.
- 1875: This year marks the establishment of the Bombay Stock Exchange.
- 1920: This year is not the establishment year of the BSE.
- 1960: This year is not the establishment year of the BSE.

Historical records show that the Bombay Stock Exchange, originally known as 'The Native Share & Stock Brokers' Association', was founded in 1875. It was the first stock exchange in Asia.

Therefore, the correct year of establishment for the Bombay Stock Exchange is 1875.

Revision Table: Key Information about BSE

Aspect	Detail
Full Name	Bombay Stock Exchange
Abbreviation	BSE
Original Name	The Native Share & Stock Brokers' Association
Establishment Year	1875
Location	Mumbai, India
Significance	Asia's first stock exchange

Additional Information on Stock Exchanges

A stock exchange is a market where buyers and sellers trade shares of publicly listed companies. They play a crucial role in the economy by:

- Providing a platform for companies to raise capital through the issue of stocks.
- Offering investors a place to buy and sell securities.
- Facilitating price discovery based on supply and demand.
- Ensuring transparency and regulation in trading activities.

The BSE, along with the National Stock Exchange (NSE), are the primary stock exchanges in India. Understanding the history of institutions like the BSE helps in grasping the evolution of India's financial markets.

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28. Answer: a

Explanation:

Understanding the 'Startup of the Year 2017' Award

The 'Startup of the Year 2017' award is a recognition given to emerging businesses that demonstrate significant potential, innovation, and growth within their respective sectors. This particular award was presented at the 7th edition of the Small Business Awards, an event aimed at celebrating achievements and excellence in the small business ecosystem in India.

The 7th Small Business Awards ceremony was held in New Delhi. These awards serve as a platform to highlight the contributions of small and medium enterprises (SMEs) and startups to the Indian economy.

Identifying the 'Startup of the Year 2017' Award Winner

Among several promising contenders, one company was specifically recognized as the 'Startup of the Year 2017' at this prestigious event in New Delhi. The recipient demonstrated innovation in its business model and execution.

Analysis of the Options

Let's look at the options provided:

- Milkbasket: Known for its daily micro-delivery service for groceries and milk.
- Milkmaid: A brand of condensed milk, part of a larger food conglomerate.
- Milkdairy: A general term related to milk production or processing.
- Milbar: This option does not represent a widely known startup or business in this context.

Based on the recognition received at the 7th Small Business Awards held in New Delhi, the company conferred with the 'Startup of the Year 2017' award was Milkbasket.

Conclusion on the Startup Award

The 'Startup of the Year 2017' award at the 7th Small Business Awards in New Delhi recognized Milkbasket for its innovative approach in the daily essential delivery space. This award highlights the company's early success and potential in the Indian market.

Revision Table: Startup of the Year 2017

Award	Year	Event	Location	Winner
Startup of the Year	2017	7th Small Business Awards	New Delhi	Milkbasket

Additional Information: Indian Startup Ecosystem & Awards

The Indian startup ecosystem has seen significant growth over the past decade. Various awards and recognition programs play a crucial role in motivating entrepreneurs and bringing attention to innovative businesses. Events like the Small Business Awards provide a platform for startups to gain visibility and connect with investors and mentors. The recognition as 'Startup of the Year' is a significant milestone for any young company, validating its business model and potential for future growth. Milkbasket, the recipient in 2017, operates in the hyper-local delivery sector, which was experiencing rapid expansion during that period, particularly for daily needs like milk and groceries.

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29. **Answer: b**

Explanation:

Understanding Direction and Distance Problems

This problem involves finding the final position relative to the starting point after a series of movements in different directions. We need to determine both the straight-line distance and the direction from the initial location.

Step-by-Step Movement Analysis

Let's break down Raju's movement:

1. Raju starts at an initial position (let's call it Point A).
2. He walks 4 km towards the East from Point A to a new point (let's call it Point B).

3. From Point B, he turns left. When facing East, a left turn is towards the North.
4. He walks another 3 km towards the North from Point B to his final position (let's call it Point C).

We now need to find the distance between his initial position (Point A) and his final position (Point C), and the direction of Point C from Point A.

Visualizing the Path: Forming a Right Triangle

The path taken by Raju forms a shape. He first moved East (AB) and then North (BC). The initial position (A), the point after the first turn (B), and the final position (C) form a right-angled triangle, with the right angle at Point B.

- Side AB = 4 km (East)
- Side BC = 3 km (North)
- The distance from the initial position to the final position is the hypotenuse of this right-angled triangle (AC).

Calculating the Distance using the Pythagorean Theorem

In a right-angled triangle, the square of the hypotenuse is equal to the sum of the squares of the other two sides. This is known as the Pythagorean theorem.

Let AC be the distance from the initial position. According to the Pythagorean theorem:

$$AC^2 = AB^2 + BC^2$$

Substitute the distances Raju walked:

$$AC^2 = (4\,\text{km})^2 + (3\,\text{km})^2$$

$$AC^2 = 16\,\text{km}^2 + 9\,\text{km}^2$$

$$AC^2 = 25\,\text{km}^2$$

To find AC, we take the square root of both sides:

$$AC = \sqrt{25\,\text{km}^2}$$

$AC = 5\,\text{km}$

So, the distance from the initial position is 5 km.

Determining the Direction from Initial Position

Now we need to find the direction of the final position (Point C) relative to the initial position (Point A).

Raju first moved East from A, and then North from B to reach C.

- Starting from A, moving East puts him along the positive x-axis (if we consider A as the origin).
- Then moving North puts him along the positive y-axis relative to the East direction.

The final position C is located both to the East and to the North of the initial position A. The direction that is both East and North is called North-East.

Summary of Findings

By calculating the hypotenuse of the right triangle formed by his movements, we found the distance. By observing the final position's location relative to the start (Eastward and Northward), we determined the direction.

Movement	Direction	Distance	Points
First Leg	East	4 km	A to B
Second Leg	North (Left Turn)	3 km	B to C
Final Position from Initial	North-East	5 km	A to C

Therefore, Raju is 5 km away from his initial position in the North-East direction.

Revision Table: Key Concepts in Direction and Distance

Concept	Description	Application in this Problem
Cardinal Directions	North, South, East, West	Used to define movement along straight lines.
Intermediate Directions	North-East, North-West, South-East, South-West	Used to define position or movement that is a combination of two cardinal directions.
Turning Left/Right	Changing direction by 90 degrees. Left turn from East is North; from North is West, etc.	Turning left from East led Raju to move North.
Displacement vs. Distance	Displacement is the straight-line distance and direction from start to end; Distance is the total path length travelled.	We calculated the displacement (5 km North-East), not the total distance travelled (4 km + 3 km = 7 km).
Pythagorean Theorem	$a^2 + b^2 = c^2$ for a right triangle, where c is the hypotenuse.	Used to calculate the straight-line distance (hypotenuse) between the start and end points.

Additional Information on Direction and Distance Reasoning

Direction and distance problems often involve visualizing movements on a 2D plane. It's helpful to imagine or sketch a coordinate system where the starting point is the origin (0,0).

- Moving East increases the x-coordinate.
- Moving West decreases the x-coordinate.
- Moving North increases the y-coordinate.
- Moving South decreases the y-coordinate.

In this problem:

- Start at A (0,0).
- Walk 4 km East: reaches B (4,0).
- Turn left (North) and walk 3 km: from B (4,0), move 3 units North to reach C (4,3).

The final position is (4,3). The distance from the origin (0,0) to (4,3) is calculated using the distance formula, which is derived from the Pythagorean theorem: $\text{Distance} = \sqrt{(x_2 - x_1)^2 + (y_2 - y_1)^2}$. Here, $\sqrt{(4-0)^2 + (3-0)^2} = \sqrt{4^2 + 3^2} = \sqrt{16+9} = \sqrt{25} = 5$ km.

The point (4,3) is in the first quadrant of a standard coordinate system where East is the positive x-axis and North is the positive y-axis. Points in the first quadrant are North-East of the origin.

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30. Answer: a

Explanation:

Understanding the Mean and Median in Statistical Data

The question asks us to find the new mean of 21 observations after a specific change is applied only to observations greater than the median. Let's break down the steps to solve this problem.

Initial State Analysis

- We are given 21 distinct observations. Let the number of observations be $n = 21$.
- The initial mean of these observations is given as 40.
- The sum of the initial observations ($\Sigma x_{initial}$) can be calculated using the formula: Mean = $\frac{\text{Sum of observations}}{\text{Number of observations}}$.

So, the initial sum of observations is:

$$\Sigma x_{initial} = \text{Mean} \times n = 40 \times 21$$

$\Sigma x_{initial} = 840$

Identifying Observations Greater Than the Median

For a set of n distinct observations arranged in ascending order, the median is the middle value. Since $n = 21$ is an odd number, the median is the observation at the $(\frac{n+1}{2})$th position.

Position of median = $\frac{21+1}{2}$th position = $\frac{22}{2}$th position = 11th position.

The observations are distinct, so when sorted, the median is the 11th value. Observations greater than the median are those values that come *after* the 11th position in the sorted list.

The positions of observations greater than the median are 12th, 13th, 14th, ..., 21st.

Number of observations greater than the median = (Last position) - (Position of median) = 21 - 11 = 10 observations.

So, there are 10 observations that are greater than the median.

Applying the Change to Observations

The problem states that each observation greater than the median is increased by 21.

- There are 10 such observations.
- Each of these 10 observations is increased by 21.

The total increase in the sum of observations is the sum of the increases applied to these 10 observations.

Total increase in sum = Number of observations increased $\times$ Amount of increase per observation

Total increase in sum = $10 \times 21 = 210$

Calculating the New Mean

The new sum of observations (Σx_{new}) is the initial sum plus the total increase in sum.

$$\Sigma x_{new} = \Sigma x_{initial} + \text{Total increase in sum}$$

$$\Sigma x_{new} = 840 + 210 = 1050$$

The number of observations remains the same, $n = 21$.

The new mean is calculated by dividing the new sum by the total number of observations.

New Mean = $\frac{\Sigma x_{new}}{n}$

New Mean = $\frac{1050}{21}$

Calculation

To calculate $\frac{1050}{21}$, we can perform the division:

$$\frac{1050}{21} = \frac{105 \times 10}{21}$$

Since $105 = 21 \times 5$, we have:

$$\frac{(21 \times 5) \times 10}{21} = 5 \times 10 = 50$$

The new mean of the observations is 50.

Summary of Calculation Steps

Step	Description	Value
1	Number of observations (n)	21
2	Initial Mean	40
3	Initial Sum (40×21)	840
4	Position of Median $\left(\frac{21+1}{2}\right)$	11th
5	Number of observations > Median (21 - 11)	10
6	Increase per observation > Median	21
7	Total increase in Sum (10×21)	210
8	New Sum $(840 + 210)$	1050
9	New Mean $\left(\frac{1050}{21}\right)$	50

The new mean of the observations is 50.

Revision Table: Mean and Median Concepts

Concept	Definition	Calculation Note
Mean	The average of a set of numbers.	Sum of observations / Number of observations
Median (Odd n)	The middle value in a sorted dataset.	Value at the $\frac{n+1}{2}$th position when sorted.
Median (Even n)	The average of the two middle values in a sorted dataset.	Average of values at the $\frac{n}{2}$th and $\frac{n}{2}+1$th positions when sorted.
Effect of adding constant to observations	If a constant c is added to EACH observation, the mean increases by c.	New Mean = Old Mean + c
Effect of multiplying constant to observations	If EACH observation is multiplied by a constant k, the mean is multiplied by k.	New Mean = Old Mean $\times k$

Additional Information on Data Transformations and Mean

This problem demonstrates how changes to a subset of data affect the overall mean. When a constant value is added to or subtracted from *some* observations, the total sum changes by the sum of these additions/subtractions. The mean then changes proportionally to this total change divided by the number of observations.

- If we increased *all* 21 observations by 21, the new mean would simply be $40+21=61$.
- However, only 10 observations were increased. This selective change means we must calculate the total change in sum and then find the new mean.
- The median itself does not directly participate in the calculation of the mean, but its position helps us identify which observations are affected by the change. The median's value might or might not change depending on the magnitude of the increases, but that is not required to solve this problem.

31. **Answer: d**

Explanation:

Understanding the Banga Bibhushan Award

The question asks about the recipient of the Banga Bibhushan award for the year 2018. The Banga Bibhushan is recognised as the highest civilian honour bestowed by the Government of West Bengal, India. This award is given to individuals for their outstanding contributions in various fields, including arts, science, literature, sports, social service, and other areas.

Recipient of Banga Bibhushan 2018

For the year 2018, the Banga Bibhushan award was conferred upon several distinguished personalities. Among the options provided, one individual was a recipient in 2018.

Based on the information available, the renowned playback singer Asha Bhosle was one of the recipients of the Banga Bibhushan award in 2018. Her significant contribution to music was recognised with this prestigious honour from the state of West Bengal.

Analysing the Options

Let's look at the provided options in the context of the Banga Bibhushan award for 2018:

- **Aparna Sen:** Aparna Sen is a highly respected film director and actress. While she has received many accolades, she was not among the recipients of the Banga Bibhushan in 2018.
- **Victor Banerjee:** Victor Banerjee is a well-known actor. Like Aparna Sen, he is a notable personality but did not receive the Banga Bibhushan in 2018.
- **Soumitra Chatterjee:** Soumitra Chatterjee was a legendary actor. He was awarded the Banga Bibhushan in an earlier year, 2012. He was not a recipient in

2018.

- **Asha Bhosle:** Asha Bhosle, the iconic singer, was indeed awarded the Banga Bibhushan in 2018 for her immense contribution to the field of music over several decades.

Therefore, the correct individual among the options who received the Banga Bibhushan, the highest civilian award of West Bengal, for 2018 is Asha Bhosle.

Revision Table: Banga Bibhushan Award

Award	State	Significance	Recipient (2018, from options)
Banga Bibhushan	West Bengal	Highest Civilian Honour	Asha Bhosle

Additional Information on West Bengal Awards

The Banga Bibhushan is one of the state honours given by West Bengal. Alongside Banga Bibhushan, another significant award is the Banga Bhushan, also recognising contributions across various fields but considered a slightly lower tier honour compared to the Banga Bibhushan.

These awards are typically announced and presented around the state's Foundation Day or another significant state occasion. The list of awardees often includes prominent figures from Bengali culture, arts, and public life, as well as individuals from outside the state who have contributed significantly to areas relevant to West Bengal or its people.

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32. **Answer: a**

Explanation:

Concept:

When the midpoints of the two sides are joined by a line then that line is parallel to the third side of the triangle.

The ratio of the area of the triangle formed by the line joining the midpoints to the area of the original triangle is equal to the ratio of the square of the corresponding sides of the triangles.

Calculation:

Here, we need not to solve whole question.

By, properties of triangle

If a new triangle is formed by joining mid points of sides of a given triangle, then ratio of area of new triangle to area of given triangle is equal to 1 : 4.

Area of given triangle = Δ 1

Area of new triangle formed by midpoints = Δ 2

$\frac{\Delta_2}{\Delta_1} = \frac{1}{4}$

Hence, $\frac{\Delta_2}{\Delta_1} = \frac{1}{x} = \frac{1}{4}$

∴ x = 4.

---

33. **Answer: c**

Explanation:

Let's break down this work and time problem step-by-step. We need to find out how many days Anjali takes to finish the remaining task after Ranjith leaves.

Understanding Work Rates

The key to solving work and time problems is to determine the individual work rate of each person. The work rate is the amount of work done by a person in one day.

- Ranjith can complete the task in 25 days. So, Ranjith's daily work rate is $\frac{1}{25}$ of the task.
- Anjali can finish the task in 20 days. So, Anjali's daily work rate is $\frac{1}{20}$ of the task.

Work Done When Working Together

Ranjith and Anjali work together for 5 days. First, let's find their combined daily work rate.

Combined daily work rate = Ranjith's daily rate + Anjali's daily rate

Combined daily work rate = $\frac{1}{25} + \frac{1}{20}$

To add these fractions, we find a common denominator, which is 100.

Combined daily work rate = $\frac{1\times4}{25\times4} + \frac{1\times5}{20\times5} = \frac{4}{100} + \frac{5}{100} = \frac{9}{100}$ of the task per day.

Now, we calculate the work they complete together in 5 days.

Work done in 5 days = Combined daily work rate $\times$ Number of days

Work done in 5 days = $\frac{9}{100} \times 5 = \frac{45}{100} = \frac{9}{20}$ of the task.

Calculating the Remaining Work

The total task is considered as 1 unit of work. After 5 days, $\frac{9}{20}$ of the task is completed.

Remaining work = Total work - Work done in 5 days

Remaining work = $1 - \frac{9}{20}$

To subtract, we write 1 as $\frac{20}{20}$.

Remaining work = $\frac{20}{20} - \frac{9}{20} = \frac{11}{20}$ of the task.

Time for Anjali to Finish Remaining Work

After 5 days, Ranjith leaves. Anjali needs to finish the remaining $\frac{11}{20}$ of the task by herself. We know Anjali's daily work rate is $\frac{1}{20}$ of the task per day.

Time taken by Anjali to finish remaining work = $\frac{\text{Remaining work}}{\text{Anjali's daily work rate}}$

Time taken by Anjali = $\frac{\frac{11}{20}}{\frac{1}{20}}$

To divide by a fraction, we multiply by its reciprocal.

Time taken by Anjali = $\frac{11}{20} \times \frac{20}{1} = \frac{11\times20}{20\times1} = 11$ days.

So, Anjali will take 11 days to finish the remaining work.

Person	Time to complete task	Daily Work Rate
Ranjith	25 days	$\frac{1}{25}$
Anjali	20 days	$\frac{1}{20}$

Activity	Calculation	Result
Combined Daily Rate	$\frac{1}{25} + \frac{1}{20} = \frac{9}{100}$	$\frac{9}{100}$ per day
Work Done in 5 days (Together)	$\frac{9}{100} \times 5$	$\frac{9}{20}$ of the task
Remaining Work	$1 - \frac{9}{20}$	$\frac{11}{20}$ of the task
Time for Anjali (Remaining)	$\frac{\frac{11}{20}}{\frac{1}{20}}$	11 days

Conclusion

After working together for 5 days and Ranjith leaving, Anjali will need 11 more days to complete the rest of the task.

Revision Table: Work and Time Concepts

Concept	Explanation	Formula
Individual Work Rate	The fraction of the total work done by a person in one unit of time (e.g., one day).	If a person completes a task in 'n' days, their daily rate is $\frac{1}{n}$.
Combined Work Rate	The sum of individual work rates when multiple people work together.	If A's rate is R_A and B's rate is R_B, combined rate is $R_A + R_B$.
Work Done	The amount of task completed over a period.	Work Done = Work Rate $\times$ Time
Time Taken	The duration required to complete a certain amount of work.	Time Taken = $\frac{\text{Amount of Work}}{\text{Work Rate}}$
Total Work	Usually considered as 1 unit or represented by the LCM of the times taken by individuals.	1 (or LCM of times)

Additional Information: Solving Work and Time Problems

Work and time problems often involve calculating how long it takes individuals or groups to complete a task. Here are some tips for solving them:

- Always find the individual's one-day work rate first. This is the reciprocal of the time taken to complete the whole task.
- If people work together, add their individual daily rates to find the combined daily rate.
- If someone leaves or joins, calculate the work done during the period they worked together, then calculate the remaining work.
- Use the remaining work and the work rate of the person(s) continuing the work to find the time needed to finish the rest.
- Be careful with units (days, hours, etc.) and ensure consistency throughout the calculation.

34. **Answer: d**

Explanation:

Understanding Bee Sting Treatment

Bee stings can be painful and cause swelling, redness, and itching. This is largely due to the venom injected by the bee, which is primarily acidic. To help alleviate the symptoms and neutralize the acidic venom, a basic or alkaline substance is often recommended for topical application.

Why Baking Soda is Used for Bee Stings

The venom of a bee sting is predominantly acidic. To counteract an acid, you need a base. Baking soda, chemically known as sodium bicarbonate ($NaHCO_3$), is a mild alkali (base). When baking soda is mixed with water to form a paste, it can be applied to the sting site. The basic properties of the baking soda help to neutralize the acidic venom, which can reduce pain, itching, and swelling.

Analysing the Options for Bee Sting Treatment

Let's look at the provided options and their properties:

- **Lemon juice:** Lemon juice is acidic, primarily containing citric acid. Applying an acid to an already acidic sting would likely worsen the reaction and increase discomfort.
- **Milk:** Milk is generally slightly acidic or close to neutral, depending on its type and freshness. It does not have strong basic properties to effectively neutralize acidic bee venom.
- **Vinegar:** Vinegar is acidic, mainly containing acetic acid. Applying an acid like vinegar to an acidic sting would be counterproductive and could increase irritation.
- **Baking soda:** Baking soda (sodium bicarbonate) is a base. It can effectively neutralize the acidic components of bee venom, providing relief from the sting symptoms.

Comparing Treatment Options

Here is a comparison of how each option relates to the acidic nature of bee sting venom:

Substance	Nature (Acid/Base)	Effect on Acidic Bee Venom
Lemon juice	Acidic	Increases acidity; counterproductive
Milk	Slightly Acidic/Neutral	Ineffective for neutralization
Vinegar	Acidic	Increases acidity; counterproductive
Baking soda	Basic	Neutralizes acidity; helpful

Based on the principle of acid-base neutralization, a basic substance is required to treat an acidic bee sting. Among the given options, only baking soda is a base.

Conclusion: The Correct Solution for Bee Sting

To treat a bee sting, a substance that can neutralize the acidic venom is needed. Baking soda is a basic substance that effectively serves this purpose. Therefore, baking soda solution is used to treat bee sting.

Revision Table: Key Concepts for Bee Stings

Concept	Description	Relevance to Bee Stings
Bee Venom	Primarily acidic mixture of compounds	Causes pain, swelling, itching upon stinging
Acid-Base Neutralization	Reaction between an acid and a base	Baking soda (base) neutralizes bee venom (acid)
Baking Soda	Sodium bicarbonate ($NaHCO_3$)	Mild alkali, used topically to neutralize acidic stings

Additional Information: Bee Sting Management

Beyond neutralizing the venom with baking soda, managing a bee sting involves several steps:

- Remove the stinger quickly if it's still present. Honeybees leave their stinger behind, while other bees and wasps do not.
- Wash the sting area with soap and water to prevent infection.
- Apply a cold compress or ice pack to reduce swelling and pain.
- Over-the-counter pain relievers like ibuprofen or acetaminophen can help with pain.
- Antihistamines can help reduce itching and swelling.
- Monitor the sting site for signs of infection or allergic reaction. Seek medical attention if symptoms worsen or if signs of a severe allergic reaction (anaphylaxis) like difficulty breathing, swelling of the face or throat, or dizziness occur.

While baking soda can help with the acidic component, proper removal of the stinger and managing swelling and allergic reactions are also crucial aspects of bee sting treatment.

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35. **Answer: c**

Explanation:

Calculating Atomic Number from Mass Number and Neutrons

This question asks us to find the atomic number of an isotope given its mass number and the number of neutrons in its nucleus. Let's break down the concepts involved.

Understanding Key Concepts

- **Atomic Number (Z):** The atomic number is the number of protons in the nucleus of an atom. This number uniquely identifies an element. For a neutral atom, the number of electrons is equal to the number of protons.
- **Mass Number (A):** The mass number is the total number of protons and neutrons in the nucleus of an atom. It is an integer.
- **Neutron Number (N):** The neutron number is the number of neutrons in the nucleus of an atom.

The relationship between these three quantities is fundamental in nuclear physics and chemistry. The mass number is simply the sum of the atomic number and the number of neutrons.

Mathematically, this relationship is expressed as:

$$A = Z + N$$

Where:

- A is the Mass Number
- Z is the Atomic Number (Number of Protons)
- N is the Number of Neutrons

Solving the Problem

We are given the following information for the isotope:

- Mass Number (A) = 298
- Number of Neutrons (N) = 189

We need to find the Atomic Number (Z). We can rearrange the formula $A = Z + N$ to solve for Z:

$$Z = A - N$$

Now, we can substitute the given values into this rearranged formula:

$$Z = 298 - 189$$

Performing the subtraction:

$Z = 109$

So, the atomic number of the isotope is 109.

Step-by-Step Calculation

1. Identify the given values: Mass Number (A) = 298, Number of Neutrons (N) = 189.
2. Recall or state the formula relating Mass Number, Atomic Number, and Neutron Number: $A = Z + N$.
3. Rearrange the formula to solve for the Atomic Number (Z): $Z = A - N$.
4. Substitute the given values into the rearranged formula: $Z = 298 - 189$.
5. Calculate the result: $Z = 109$.

The atomic number is 109. This atomic number corresponds to the element Meitnerium (Mt).

Revision Table: Nuclear Properties

Property	Symbol	Definition	Calculation
Atomic Number	Z	Number of Protons	$Z = A - N$
Mass Number	A	Protons + Neutrons	$A = Z + N$
Neutron Number	N	Number of Neutrons	$N = A - Z$

Additional Information: Isotopes and Elements

An isotope is an atom of an element that has the same number of protons (and thus the same atomic number, Z) but a different number of neutrons (N). Because the number of neutrons differs, the mass number $(A = Z + N)$ also differs between isotopes of the same element.

- The atomic number (Z) is what defines an element. All atoms with $Z = 6$ are carbon atoms, regardless of their neutron count.

- Isotopes of an element have the same chemical properties because chemical properties are primarily determined by the number of electrons, which in a neutral atom is equal to the number of protons (Z).
- Isotopes differ in their physical properties, such as mass and density, and can have different nuclear stabilities.

In this problem, the atomic number is 109. Any atom with 109 protons is an atom of the element with atomic number 109 (Meitnerium). The fact that it has a mass number of 298 means it is a specific isotope of Meitnerium, specifically, it is ^{298}Mt.

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36. **Answer: b**

Explanation:

There are four symbols in each figure:

Let us number each position from left to right from 1 to 4.

The symbol **'⊠'** moves one position to the right. So, in the figure 4 it will be at position 2.

The symbol **'>'** moves one position to the right. So, in the figure 4 it will be at position 3.

The places that are left are 1 and 4.

The symbol **'*'** changes their position around **'Δ'**

In figure 1 star is to the left of triangle.

In figure 2 star is to the right of triangle.

In figure 3 star is to the left of triangle.

So, in the figure 4 star will be at position 4.

Hence, the figure 4 will be figure D:

Hence, D is the correct answer.

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37. **Answer: c**

Explanation:

Understanding LCM of Co-prime Numbers

The question asks for the Least Common Multiple (LCM) of two co-prime numbers, denoted as 'a' and 'b', where 'a' is greater than 'b'. To answer this, we first need to understand what co-prime numbers are and how LCM is calculated.

What are Co-prime Numbers?

Two numbers are considered co-prime (or relatively prime) if their only common positive factor or divisor is 1. This means their Greatest Common Divisor (GCD) is 1.

- Example: The numbers 7 and 10 are co-prime. The factors of 7 are 1 and 7. The factors of 10 are 1, 2, 5, and 10. The only common factor is 1. So, GCD(7, 10) = 1.
- Example: The numbers 6 and 9 are not co-prime. The factors of 6 are 1, 2, 3, 6. The factors of 9 are 1, 3, 9. The common factors are 1 and 3. The GCD(6, 9) = 3.

Calculating the Least Common Multiple (LCM)

The LCM of two or more numbers is the smallest positive integer that is a multiple of all the numbers. There are several ways to find the LCM, but one common method uses the relationship between LCM, GCD, and the product of the two numbers. For any two positive integers 'a' and 'b', the following relationship holds:

$$\text{LCM}(a, b) \times \text{GCD}(a, b) = a \times b$$

Finding LCM for Co-prime Numbers

Now, let's apply this relationship to co-prime numbers. By definition, if 'a' and 'b' are co-prime numbers, their Greatest Common Divisor (GCD) is 1.

So, $\text{GCD}(a, b) = 1$.

Substitute this into the relationship formula:

$$\text{LCM}(a, b) \times 1 = a \times b$$

This simplifies to:

$$\text{LCM}(a, b) = a \times b$$

Therefore, the LCM of two co-prime numbers is simply their product.

Example Calculation

Let's take the co-prime numbers 7 and 10 again. We know they are co-prime, so GCD(7, 10) = 1.

Using the formula, LCM(7, 10) = 7 $\times$ 10 = 70.

Let's verify by listing multiples:

- Multiples of 7: 7, 14, 21, 28, 35, 42, 49, 56, 63, **70**, ...
- Multiples of 10: 10, 20, 30, 40, 50, 60, **70**, ...

The smallest common multiple is indeed 70, which is the product of 7 and 10.

Conclusion

For any two co-prime numbers 'a' and 'b', their Greatest Common Divisor (GCD) is 1. Using the fundamental relationship $\text{LCM}(a, b) \times \text{GCD}(a, b) = a \times b$, we find that $\text{LCM}(a, b) \times 1 = a \times b$, which means $\text{LCM}(a, b) = a \times b$. The condition that 'a' is greater than 'b' does not affect the LCM calculation.

Summary: Co-prime Numbers and LCM

Concept	Definition	Relationship for Co-prime a, b
Co-prime Numbers	Two numbers with GCD = 1	GCD(a, b) = 1
LCM (Least Common Multiple)	Smallest common multiple	LCM(a, b) = a $\times$ b
LCM-GCD Relation	LCM(a, b) $\times$ GCD(a, b) = a $\times$ b	(a $\times$ b) $\times$ 1 = a $\times$ b (Confirms LCM = ab)

Revision Table: Number Theory Concepts

Key Definitions in Number Theory

Term	Definition	Example
Factor (Divisor)	A number that divides another number exactly.	Factors of 12 are 1, 2, 3, 4, 6, 12.
Multiple	A number that can be divided by another number without a remainder.	Multiples of 5 are 5, 10, 15, 20, ...
GCD (Greatest Common Divisor)	The largest positive integer that divides two or more numbers without a remainder. Also called HCF (Highest Common Factor).	GCD(12, 18) = 6.
LCM (Least Common Multiple)	The smallest positive integer that is a multiple of two or more numbers.	LCM(12, 18) = 36.
Co-prime Numbers	Two numbers whose only common positive factor is 1 (i.e., their GCD is 1).	7 and 15 are co-prime because GCD(7, 15) = 1.

Additional Information: Properties of Co-prime Numbers and LCM

- Two prime numbers are always co-prime. For example, 3 and 7 are prime, and GCD(3, 7) = 1. Their LCM is 3 $\times$ 7 = 21.
- A prime number and any other number that is not a multiple of the prime number are co-prime. For example, 5 (prime) and 12 (not a multiple of 5). GCD(5, 12) = 1. LCM(5, 12) = 5 $\times$ 12 = 60.
- If two numbers are co-prime, their ratio is in its simplest form (lowest terms).
- The property $\text{LCM}(a, b) = a \times b$ for co-prime numbers is a direct consequence of the relationship $\text{LCM}(a, b) \times \text{GCD}(a, b) = a \times b$. Since GCD is 1, the product is the LCM.
- The concept of co-prime numbers is fundamental in various areas of mathematics, including number theory, cryptography, and fractions.

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38. **Answer: a**

Explanation:

Understanding the Question: Arranging by Height

The question asks us to determine who is standing last when four people (M, A, N, K) are arranged in a row from shortest to tallest. This means we need to identify the tallest person among the four.

We are given two statements and need to evaluate if they, individually or together, provide enough information to find the tallest person.

Analyzing Statement 1: A is Shorter Than K

Statement 1 provides a height comparison between A and K:

- Statement 1: A is shorter than K $(\text{A} < \text{K})$

This tells us that K is taller than A. However, we have no information about M and N relative to A or K. We cannot determine the tallest person among all four using only this statement.

Conclusion for Statement 1 alone: Not sufficient.

Analyzing Statement 2: M is Shorter Than A

Statement 2 provides a height comparison between M and A:

- Statement 2: M is shorter than A $(\mathrm{M} < \mathrm{A})$

This tells us that A is taller than M. Again, we have no information about K and N relative to M or A. We cannot determine the tallest person among all four using only this statement.

Conclusion for Statement 2 alone: Not sufficient.

Analyzing Statements 1 and 2 Together: Combined Information

Now, let's combine the information from both statements:

- From Statement 1: $\mathrm{A} < \mathrm{K}$
- From Statement 2: $\mathrm{M} < \mathrm{A}$

Combining these inequalities, we can establish a height order for M, A, and K:

$$\mathrm{M} < \mathrm{A} < \mathrm{K}$$

This sequence shows that among M, A, and K, M is the shortest, A is in the middle, and K is the tallest. However, we still have no information about the height of person N.

N's height could be anywhere relative to M, A, and K:

- N could be shorter than M $(\mathrm{N} < \mathrm{M})$. In this case, K would be the tallest.
- N could be between M and A $(\mathrm{M} < \mathrm{N} < \mathrm{A})$. K would still be the tallest.
- N could be between A and K $(\mathrm{A} < \mathrm{N} < \mathrm{K})$. K would still be the tallest.

- N could be taller than K ($N > K$). In this case, N would be the tallest.

Since N's height relative to K is unknown, we cannot definitively say whether K is the tallest person overall or if N is the tallest. Therefore, we cannot determine who is standing last (the tallest person) when arranged from shortest to tallest using statements 1 and 2 together.

Conclusion for Statements 1 and 2 together: Not sufficient.

Summary of Sufficiency Analysis

Condition	Sufficient to answer?	Reasoning
Statement 1 alone	No	Only $A < K$ is known. Information about M and N is missing for the height arrangement.
Statement 2 alone	No	Only $M < A$ is known. Information about K and N is missing for the height arrangement.
Statements 1 and 2 together	No	Known order is $M < A < K$. Information about N's height relative to K is missing. N could be taller than K, affecting the final height arrangement.

Based on the analysis of the height arrangement information provided by the statements, neither statement alone nor both statements together provide enough information to uniquely identify the tallest person among M, A, N, and K.

Final Answer Determination for the Height Arrangement

To answer the question "who is standing last?" (meaning the tallest person), we need to know the height of all four individuals relative to each other. The combined statements only establish the relative heights of M, A, and K but give no information about N's height or its position in the height arrangement. Therefore, we cannot determine the tallest person.

Revision Table: Data Sufficiency Concepts for Height Arrangement

Concept	Explanation	Application in this Height Arrangement Problem
Data Sufficiency	Evaluating if given facts are enough to solve a specific problem or answer a question.	Determining if statements about relative heights are sufficient to find the tallest person for the height arrangement.
Sufficiency	A set of facts is sufficient if it leads to one definite answer.	Statements are not sufficient because N's position relative to K is unknown in the height arrangement, meaning the tallest person cannot be definitively identified.
Necessary	A statement is essential for solving the problem. (Contextual with sufficiency).	While the statements provide some relative height data for the arrangement, they are not sufficient together due to the missing information about N.

Additional Information: Solving Ordering and Height Arrangement Problems

Problems involving ordering or ranking individuals based on criteria like height, age, marks, etc., often require establishing the relative position of everyone in the set for the final arrangement. Here are some tips:

- **Visualize the Order:** Use symbols like $<$ or $>$ to represent the relationships (e.g., $A < K$ for A is shorter than K).
- **Combine Relational Statements:** Look for connections to build longer sequences. If $M < A$ and $A < K$, combine to $M < A < K$.
- **Identify Unknown Relationships:** Note which individuals' relative positions are not defined by the statements. This is crucial for checking sufficiency for the

height arrangement.

- **Check for Uniqueness:** If the combined information allows for more than one possible final arrangement or answer to the question, the statements are not sufficient.
- **Missing Data:** A problem with missing data about one or more elements in the set usually results in insufficient statements unless the known data can definitively answer the specific question asked regardless of the missing data.

In this specific height arrangement problem, the lack of information about N's height relative to K means we cannot be sure who is the tallest, proving that the given statements are not sufficient.

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39. **Answer: d**

Explanation:

Understanding Constantan Alloy Composition

Constantan is a well-known alloy widely used in applications requiring stable electrical resistance over a range of temperatures. Understanding its composition is fundamental to knowing its properties and uses.

The question asks about the metallic components from which Constantan is alloyed. An alloy is a mixture of two or more elements, where at least one is a metal. Constantan is specifically an alloy of copper and nickel.

Components of Constantan

Constantan is primarily composed of:

- Copper (Cu)
- Nickel (Ni)

Typical composition varies slightly, but it is generally about 55% copper and 45% nickel. Small amounts of other elements like manganese (Mn) or iron (Fe) might sometimes be present as impurities or minor additions, but the defining components are copper and nickel.

Analyzing the Options

Let's look at the given options and compare them to the known composition of Constantan:

Option	Components	Is it Constantan?	Reason
1	Ni, Ti, Fe, Cr	No	This combination includes Titanium (Ti), Iron (Fe), and Chromium (Cr) but misses Copper (Cu), which is a primary component of Constantan. This mix resembles stainless steel or other nickel-based alloys.
2	Cu, Ni, Mn	Possibly, but not the primary definition	This includes Copper (Cu) and Nickel (Ni), the main components. While Manganese (Mn) can sometimes be present in small amounts, Constantan is defined by the Cu-Ni binary alloy. This option is closer than 1 or 3, but option 4 is the most accurate primary definition.
3	Ni, Cr, Mn, Fe	No	Similar to option 1, this includes Nickel (Ni), Chromium (Cr), Manganese (Mn), and Iron (Fe) but lacks the essential Copper (Cu). This combination is typical of Nichrome or stainless steel variants, not Constantan.
4	Cu, Ni	Yes	This option correctly identifies Copper (Cu) and Nickel (Ni) as the primary metals alloyed to form Constantan. This is the core composition responsible for its characteristic electrical properties.

Based on the analysis, the alloy Constantan is primarily composed of Copper (Cu) and Nickel (Ni).

Revision Table: Key Facts about Constantan

Property	Description
Composition	Approximately 55% Copper (Cu), 45% Nickel (Ni)
Type	Metal alloy
Key Property	Nearly constant electrical resistance over a wide temperature range
Temperature Coefficient of Resistivity	Very low
Applications	Thermocouples (with Copper or Iron), strain gauges, resistors, heating elements
Appearance	Silvery white

Additional Information on Alloys and Materials

Understanding alloys like Constantan helps us appreciate how mixing metals can create materials with tailored properties. Here's some more information:

- **What is an Alloy?** An alloy is a mixture of chemical elements, at least one of which is a metal, to form a new substance with properties that are often different from the original elements. Alloying is done to improve strength, hardness, corrosion resistance, electrical properties, or melting point.
- **Why Alloy Copper and Nickel?** Copper is highly conductive but its resistance changes significantly with temperature. Nickel has higher resistance. Alloying them in specific proportions (like Constantan) results in a material with relatively high resistance and a very low-temperature coefficient of resistance, making it ideal for precision resistors and thermocouples where stable resistance or predictable voltage generation over temperature is needed.
- **Other Common Alloys:**
    - Brass: Copper and Zinc
    - Bronze: Copper and Tin
    - Stainless Steel: Iron, Chromium, Nickel (and sometimes Carbon, Manganese, Silicon)

- Nichrome: Nickel and Chromium

Constantan's specific Cu-Ni composition gives it unique electrical properties that are crucial for many scientific and industrial instruments.

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40. **Answer: b**

Explanation:

Understanding the Age Problem

This question involves a word problem about the ages of a father and his son at different points in time. We are given their present ages and need to determine the father's age relative to the son's present age at the time the son was born.

Given Information:

- Father's present age = 48 years
- Son's present age = 28.8 years

Calculating the Son's Birth Time

The son was born 28.8 years ago. To find the father's age when the son was born, we need to subtract the number of years that have passed since the son's birth from the father's present age.

Calculating Father's Age at Son's Birth

Father's age when son was born = Father's present age - Years passed since son's birth

Years passed since son's birth = Son's present age = 28.8 years

So, Father's age when son was born = $48 - 28.8$ years

Calculation:

$$\begin{array}{r} 48.0 \\ -28.8 \\ \hline 19.2 \end{array}$$

The father was 19.2 years old when his son was born.

Relating Father's Past Age to Son's Present Age

The question asks the father to express his age when the son was born (19.2 years) as a fraction of the son's *present* age (28.8 years). We need to find the value of X such that:

Father's age when son was born = $X \times$ Son's present age

$19.2 = X \times 28.8$

To find X, we rearrange the equation:

$X = \frac{19.2}{28.8}$

Simplifying the Fraction

To simplify the fraction $\frac{19.2}{28.8}$, we can remove the decimal points by multiplying both the numerator and denominator by 10:

$X = \frac{192}{288}$

Now, we find the greatest common divisor (GCD) of 192 and 288 or simplify by dividing by common factors. Let's divide by small factors first:

- Divide by 2: $\frac{192 \div 2}{288 \div 2} = \frac{96}{144}$
- Divide by 2 again: $\frac{96 \div 2}{144 \div 2} = \frac{48}{72}$
- Divide by 2 again: $\frac{48 \div 2}{72 \div 2} = \frac{24}{36}$
- Divide by 12: $\frac{24 \div 12}{36 \div 12} = \frac{2}{3}$

So, $X = \frac{2}{3}$.

Conclusion

The father was $\frac{2}{3}$ (two-third) of his son's present age when the son was born. Therefore, the father can say to his son, "I was two-third of your present age when you were born".

Matching with Options

- Option 1: Half $(\frac{1}{2})$ - Incorrect
- Option 2: Two-third $(\frac{2}{3})$ - Correct
- Option 3: One-fifth $(\frac{1}{5})$ - Incorrect
- Option 4: One-third $(\frac{1}{3})$ - Incorrect

Revision Table: Age Problem Summary

Concept	Value
Son's Present Age	28.8 years
Father's Present Age	48 years
Time Passed Since Son's Birth	28.8 years
Father's Age When Son Was Born	$48 - 28.8 = 19.2$ years
Ratio (Father's past age / Son's present age)	$\frac{19.2}{28.8} = \frac{2}{3}$

Additional Information on Age Calculation Problems

Age calculation problems are common in mathematics and reasoning tests. They often involve relating ages at different points in time. Here are some key ideas:

- **Time Difference:** The difference in age between two people remains constant throughout their lives. If person A is 5 years older than person B today, they will always be 5 years older, whether it was 10 years ago or will be 20 years from now.
- **Past/Future Ages:** To find someone's age in the past, subtract the number of years from their present age. To find someone's age in the future, add the number of years to their present age.

- **Setting up Equations:** For more complex problems, it is helpful to assign variables (like x or y) to the current ages and set up equations based on the information given about their ages in the past or future.
- **Fractions and Ratios:** Many problems involve expressing one age as a fraction or ratio of another age, often at a different point in time, as seen in this problem. Simplifying fractions is a key skill here.

Practice with different types of age problems helps build confidence in solving them accurately.

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41. **Answer: d**

Explanation:

Fishes are aquatic vertebrates that possess a unique circulatory system adapted to their environment. Understanding the structure of a fish heart is key to answering this question.

Understanding Fish Heart Chambers

The heart of a fish is simpler than that of terrestrial vertebrates like mammals or birds. It is located near the gills and is the central pump for their circulatory system.

A fish heart consists of the following chambers:

- An atrium (sometimes called the auricle). This chamber receives deoxygenated blood from the body.
- A ventricle. This chamber pumps the deoxygenated blood towards the gills.

These two main chambers, the atrium and the ventricle, are the functional pumping units of the fish heart.

Fish Circulatory System: Single Circulation

Fishes have a single circulation system. This means that blood passes through the heart only once in a complete circuit around the body. The path of blood flow in a fish is as follows:

1. Deoxygenated blood from the body enters the atrium.
2. From the atrium, the blood moves into the ventricle.
3. The ventricle pumps the deoxygenated blood to the gills.
4. In the gills, gas exchange occurs: carbon dioxide is released, and oxygen is taken up. The blood becomes oxygenated.
5. The oxygenated blood then flows directly from the gills throughout the rest of the body (to organs and tissues) without returning to the heart.
6. After delivering oxygen and collecting carbon dioxide in the body tissues, the deoxygenated blood returns to the heart (atrium), completing the single circuit.

This single circulation system, powered by a two-chambered heart (atrium and ventricle), is characteristic of fish.

Therefore, fishes have a **two**-chambered heart.

Revision Table: Heart Chambers in Different Vertebrates

Vertebrate Group	Number of Heart Chambers	Chamber Types	Circulation Type
Fishes	Two	One Atrium, One Ventricle	Single Circulation
Amphibians (most)	Three	Two Atria, One Ventricle	Double Circulation (Incomplete)
Reptiles (most)	Three (some have partially divided ventricle)	Two Atria, One Ventricle (partially divided)	Double Circulation (Incomplete)
Birds	Four	Two Atria, Two Ventricles	Double Circulation (Complete)
Mammals	Four	Two Atria, Two Ventricles	Double Circulation (Complete)

Additional Information on Fish Heart

While the fish heart is primarily described as having two chambers (atrium and ventricle), it also has two other structures often associated with it:

- Sinus Venosus: A thin-walled sac that receives deoxygenated blood from the body before it enters the atrium.
- Conus Arteriosus (or Bulbus Arteriosus in teleosts - bony fish): A muscular or elastic tube that receives blood from the ventricle and smooths out blood flow to the gills.

However, when referring to the main pumping chambers, the atrium and ventricle are counted, leading to the description of a two-chambered fish heart.

The efficiency of oxygen delivery in fish is facilitated by the gills and the direct flow of oxygenated blood from gills to the body, bypassing the heart on the return journey with oxygenated blood.

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42. **Answer: d**

Explanation:

Finding the Remainder When a Sum of Powers is Divided by 25

The question asks us to find the remainder when the expression $7^{21} + 7^{22} + 7^{23} + 7^{24}$ is divided by 25. This involves using concepts from modular arithmetic.

Let the given expression be S:

$$S = 7^{21} + 7^{22} + 7^{23} + 7^{24}$$

We can factor out the lowest power of 7, which is 7^{21}, from the expression:

$$S = 7^{21}(1 + 7^1 + 7^2 + 7^3)$$

Next, let's evaluate the sum inside the parentheses:

$$1 + 7^1 + 7^2 + 7^3 = 1 + 7 + 49 + 343$$

Adding these terms together:

$$1 + 7 = 8$$

$$49 + 343 = 392$$

$$8 + 392 = 400$$

So, the expression simplifies to:

$$S = 7^{21} \times 400$$

Now, we need to find the remainder when S is divided by 25. This is equivalent to finding $S \pmod{25}$:

$$S \pmod{25} = (7^{21} \times 400) \pmod{25}$$

We need to determine the remainder of 400 when divided by 25.

We can perform the division:

$$400 \div 25$$

Since $400 = 16 \times 25$, 400 is a multiple of 25.

Therefore, the remainder of 400 when divided by 25 is 0.

$$400 \equiv 0 \pmod{25}$$

Now, substitute this back into the modular expression for S:

$$S \pmod{25} = (7^{21} \times 0) \pmod{25}$$

Any number multiplied by 0 is 0. And the remainder of 0 when divided by 25 is 0.

$$S \pmod{25} = 0 \pmod{25} = 0$$

Thus, the remainder when $7^{21} + 7^{22} + 7^{23} + 7^{24}$ is divided by 25 is 0.

Step-by-Step Solution Summary

1. Identify the expression: $S = 7^{21} + 7^{22} + 7^{23} + 7^{24}$.
2. Factor out the common term 7^{21}: $S = 7^{21}(1 + 7^1 + 7^2 + 7^3)$.
3. Calculate the sum inside the parentheses: $1 + 7 + 49 + 343 = 400$.
4. Rewrite the expression: $S = 7^{21} \times 400$.
5. Determine the remainder when S is divided by 25, which is $S \pmod{25}$.
6. Find the remainder of 400 when divided by 25: $400 = 16 \times 25$, so $400 \equiv 0 \pmod{25}$.
7. Substitute this back into the expression for $S \pmod{25}$: $(7^{21} \times 0) \pmod{25}$.
8. Calculate the final remainder: $0 \pmod{25} = 0$.

Verification using Modular Properties

Alternatively, we could look at each term modulo 25. However, factoring simplifies the calculation significantly because the sum of the initial terms in the sequence $1 + 7 + 7^2 + 7^3$ resulted in a multiple of 25.

Let's check the powers of 7 modulo 25:

- $7^1 \equiv 7 \pmod{25}$
- $7^2 = 49 \equiv 24 \pmod{25}$ or $\equiv -1 \pmod{25}$
- $7^3 \equiv 7 \times (-1) = -7 \equiv 18 \pmod{25}$
- $7^4 \equiv (-1) \times (-1) = 1 \pmod{25}$

The powers of 7 modulo 25 repeat in a cycle of length 4: {7, 24, 18, 1}.

The sum inside the parentheses is $1 + 7 + 7^2 + 7^3 \equiv 1 + 7 + 24 + 18 \pmod{25}$.

$$1 + 7 + 24 + 18 = 8 + 24 + 18 = 32 + 18 = 50$$

$$50 \pmod{25} = 0$$

So, $1 + 7 + 7^2 + 7^3 \equiv 0 \pmod{25}$.

The original expression is $S = 7^{21}(1 + 7 + 7^2 + 7^3)$. Taking this modulo 25:

$$S \pmod{25} = (7^{21} \pmod{25}) \times (1 + 7 + 7^2 + 7^3 \pmod{25}) \pmod{25}$$

$$S \pmod{25} = (7^{21} \pmod{25}) \times (0) \pmod{25}$$

$$S \pmod{25} = 0$$

Both methods lead to the same result, confirming that the remainder is 0.

Revision Table: Key Concepts

Concept	Description	Relevance to Problem
Modular Arithmetic	A system of arithmetic for integers, where numbers "wrap around" upon reaching a certain value, the modulus. Finding the remainder when one number is divided by another.	The core method used to solve the problem (finding a remainder modulo 25).
Factoring Expressions	Rewriting an expression as a product of its factors. Finding a common factor in a sum.	Factoring 7^{21} simplified the sum significantly.
Remainders	The amount "left over" after division of one integer by another. Represented by $a \pmod m$.	The final answer required is the remainder when the sum is divided by 25.
Properties of Modulo	$(a \times b) \pmod m \equiv (a \pmod m) \times b \pmod m) \pmod m$. If $a \equiv 0 \pmod m$, then $a \times b \equiv 0 \pmod m$.	Used to calculate the remainder of the product $7^{21} \times 400 \pmod{25}$.

Additional Information on Modular Arithmetic and Remainders

Modular arithmetic is often called clock arithmetic because the numbers wrap around like the hours on a clock face. For example, on a 12-hour clock, 10 + 4 = 2. In modular arithmetic terms, $14 \equiv 2 \pmod{12}$.

The notation $a \equiv b \pmod m$ means that a and b have the same remainder when divided by m. Equivalently, $a - b$ is a multiple of m.

Properties used in this solution:

- If $a \equiv b \pmod{m}$, then $ac \equiv bc \pmod{m}$ for any integer c.
- If $a \equiv b \pmod{m}$ and $c \equiv d \pmod{m}$, then $a + c \equiv b + d \pmod{m}$ and $ac \equiv bd \pmod{m}$.

In this problem, finding $400 \pmod{25}$ was crucial. Since $400 = 16 \times 25$, it is perfectly divisible by 25, leaving a remainder of 0. This property, that a multiple of the modulus has a remainder of 0, is fundamental in modular arithmetic and often simplifies calculations significantly.

When dealing with powers modulo m, finding the cycle length of the powers (like we did for $7^k \pmod{25}$) can be very helpful, especially for large exponents. Euler's totient theorem provides the theoretical basis for such cycles when the base is coprime to the modulus: $a^{\phi(m)} \equiv 1 \pmod{m}$ if $\gcd(a, m) = 1$. For $m = 25$, $\phi(25) = 20$, so $7^{20} \equiv 1 \pmod{25}$.

However, in this specific problem, the sum of the first few terms in the sequence of powers $(1 + 7 + 7^2 + 7^3)$ being a multiple of 25 provided a more direct path to the solution after factoring.

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43. **Answer: b**

Explanation:

Understanding Weight and Gravity's Influence

The weight of a body is the force exerted on it due to gravity. It is calculated using the formula:

$$W = mg$$

Where:

- W is the weight of the body
- m is the mass of the body (which remains constant)
- g is the acceleration due to gravity

Since the mass of a body is constant, its weight depends entirely on the value of the acceleration due to gravity (g) at its location. The question asks where the weight is maximum, which means we need to find where the acceleration due to gravity (g) is maximum.

Factors Affecting Acceleration Due to Gravity (g) on Earth

The acceleration due to gravity on the Earth's surface is primarily influenced by two factors:

1. **Earth's Shape:** Earth is not a perfect sphere. It is an oblate spheroid, meaning it is flattened at the poles and bulges at the equator. This difference in shape means that points on the surface at the poles are slightly closer to the Earth's center of mass than points on the surface at the equator. The force of gravity follows an inverse square law, meaning it is stronger when the distance is smaller. Thus, gravity is slightly stronger at the poles due to the shorter distance to the center.
2. **Earth's Rotation:** The Earth rotates on its axis. This rotation causes a centrifugal force that acts outwards, away from the axis of rotation. This centrifugal force opposes the force of gravity, effectively reducing the apparent weight of a body. The effect of this centrifugal force is maximum at the equator, where the rotational speed is highest, and minimum (virtually zero) at the poles, where the rotational speed is zero.

Weight Variation on Earth: Poles vs. Equator

Considering both factors:

- At the **poles**, the distance to the Earth's center is minimum, and the effect of rotation is negligible. Both factors contribute to a relatively higher value of g.
- At the **equator**, the distance to the Earth's center is maximum, and the outward centrifugal force due to rotation is maximum, reducing the effective gravity. Both factors contribute to a relatively lower value of g.

Therefore, the acceleration due to gravity (g) is maximum at the poles and minimum at the equator.

Location	Distance from Earth's Center	Effect of Rotation	Resulting 'g'	Resulting Weight
Pole	Minimum	Negligible	Maximum	Maximum
Equator	Maximum	Maximum reduction	Minimum	Minimum

Analyzing the Options for Maximum Weight

Based on the variation of g, the weight of a body will be maximum where g is maximum.

1. Equator: g is minimum here.
2. Pole: g is maximum here.
3. Subtropics: Latitudes between the tropics and the temperate zones (roughly 23.5° to 40°). g is between the equatorial and polar values.
4. Tropics: Region between the Tropic of Cancer and the Tropic of Capricorn (23.5° N and 23.5° S latitude). g is relatively close to the equatorial value, though slightly higher.

Comparing these locations, the poles have the maximum value of acceleration due to gravity, and thus, the weight of a body is maximum at the poles.

The final answer is **Pole**.

Revision Table: Weight Variation Factors

Factor	How it affects 'g'	Effect at Poles	Effect at Equator
Earth's Shape (distance)	'g' decreases with distance from center	Closer to center, higher 'g'	Further from center, lower 'g'
Earth's Rotation (centrifugal force)	Reduces effective 'g'	No reduction (zero speed)	Maximum reduction (highest speed)

Additional Information: Weight vs. Mass

It is important to distinguish between weight and mass. Mass is a measure of the amount of matter in a body and is a scalar quantity. Mass is an intrinsic property of the body and remains constant regardless of location. Weight, on the other hand, is a force and is a vector quantity. It is the force of gravity acting on the mass, and its value depends on the strength of the gravitational field (g) at the location.

For example, an object with a mass of 1 kg will have a mass of 1 kg everywhere on Earth, on the Moon, or in space. However, its weight will be different:

- On Earth (average $g \approx 9.8\,\mathrm{m/s^2}$): Weight $\approx 1 \times 9.8 = 9.8\,\mathrm{N}$.
- At Earth's Poles ($g \approx 9.83\,\mathrm{m/s^2}$): Weight $\approx 1 \times 9.83 = 9.83\,\mathrm{N}$.
- At Earth's Equator ($g \approx 9.78\,\mathrm{m/s^2}$): Weight $\approx 1 \times 9.78 = 9.78\,\mathrm{N}$.
- On the Moon ($g_{\mathrm{Moon}} \approx 1.62\,\mathrm{m/s^2}$): Weight $\approx 1 \times 1.62 = 1.62\,\mathrm{N}$.

Thus, while mass is constant, weight varies with the local acceleration due to gravity.

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44. **Answer: c**

Explanation:

Understanding Recurring and Terminating Decimals

A rational number, which can be expressed as a fraction $\frac{a}{b}$ where a and b are integers and $b \neq 0$, will result in either a terminating decimal or a recurring (non-terminating repeating) decimal.

The nature of the decimal expansion depends on the prime factors of the denominator of the fraction when it is in its simplest form (reduced form). A fraction $\frac{a}{b}$ in simplest form will yield:

- A **terminating decimal** if the prime factors of the denominator b are only 2s and/or 5s.
- A **recurring decimal** if the prime factors of the denominator b include any prime numbers other than 2 or 5.

Let's analyze each given option by simplifying the fraction and examining the prime factors of the denominator.

Analyzing Each Option to Find the Recurring Decimal

Option 1: $\frac{24}{60}$

First, simplify the fraction:

$$\frac{24}{60} = \frac{12\times2}{12\times5} = \frac{2}{5}$$

The simplified denominator is 5. The prime factors of 5 are just 5. Since the only prime factor is 5, the decimal expansion will be terminating.

Decimal value: $\frac{2}{5} = 0.4$ (Terminating decimal)

Option 2: $\frac{24}{30}$

Simplify the fraction:

$$\frac{24}{30} = \frac{6\times4}{6\times5} = \frac{4}{5}$$

The simplified denominator is 5. The prime factors of 5 are just 5. Since the only prime factor is 5, the decimal expansion will be terminating.

Decimal value: $\frac{4}{5} = 0.8$ (Terminating decimal)

Option 3: $\frac{24}{90}$

Simplify the fraction:

$\frac{24}{90} = \frac{6 \times 4}{6 \times 15} = \frac{4}{15}$

The simplified denominator is 15. Find the prime factors of 15:

$15 = 3 \times 5$

The prime factors of 15 are 3 and 5. Since there is a prime factor (3) other than 2 or 5, the decimal expansion will be recurring.

Decimal value: $\frac{4}{15} = 0.2666... = 0.2\overline{6}$ (Recurring decimal)

Option 4: $\frac{24}{120}$

Simplify the fraction:

$\frac{24}{120} = \frac{24}{24 \times 5} = \frac{1}{5}$

The simplified denominator is 5. The prime factors of 5 are just 5. Since the only prime factor is 5, the decimal expansion will be terminating.

Decimal value: $\frac{1}{5} = 0.2$ (Terminating decimal)

Summary of Results

Fraction	Simplified Fraction	Simplified Denominator	Prime Factors of Denominator	Decimal Type
$\frac{24}{60}$	$\frac{2}{5}$	5	5	Terminating
$\frac{24}{30}$	$\frac{4}{5}$	5	5	Terminating
$\frac{24}{90}$	$\frac{4}{15}$	15	3, 5	Recurring
$\frac{24}{120}$	$\frac{1}{5}$	5	5	Terminating

Based on the analysis, the fraction that yields a recurring decimal is $\frac{24}{90}$, because its simplified denominator (15) has a prime factor (3) other than 2 or 5.

Revision Table: Recurring and Terminating Decimals

Decimal Type	Condition on Simplified Denominator's Prime Factors	Example
Terminating Decimal	Only 2s and/or 5s	$\frac{3}{8} = \frac{3}{2^3}$ (Denominator 8, factors: 2) $\frac{7}{20} = \frac{7}{2^2 \times 5}$ (Denominator 20, factors: 2, 5) $\frac{9}{50} = \frac{9}{2 \times 5^2}$ (Denominator 50, factors: 2, 5)
Recurring Decimal	Includes prime factors other than 2 or 5	$\frac{1}{3}$ (Denominator 3, factor: 3) $\frac{5}{6} = \frac{5}{2 \times 3}$ (Denominator 6, factors: 2, 3) $\frac{8}{11}$ (Denominator 11, factor: 11) $\frac{4}{15} = \frac{4}{3 \times 5}$ (Denominator 15, factors: 3, 5)

Additional Information on Rational Numbers and Decimal Expansion

Rational numbers are numbers that can be written as a fraction $\frac{p}{q}$, where p and q are integers and $q \neq 0$. When you convert a rational number to its decimal form, the decimal representation will always be either terminating or repeating (recurring). It will never be a non-terminating and non-repeating decimal. Numbers with non-terminating and non-repeating decimal expansions are called irrational numbers (like π or $\sqrt{2}$). The property of whether a rational number has a terminating or recurring decimal is solely determined by the prime factors of the denominator of its simplified fraction form, as discussed above.

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45. **Answer: d**

Explanation:

Union Minister for Ministry of Women and Child Development in 2018

The question asks about the individual who held the position of Union Minister for the Ministry of Women and Child Development (MWCD) in India as of the year 2018.

The Ministry of Women and Child Development is a government ministry in India that promotes the development of women and children. It was established as a separate Ministry in 2006, having previously been a department under the Ministry of Human Resource Development.

As of 2018, the Union Minister for this important ministry was actively involved in implementing policies and programs aimed at the welfare, protection, and development of women and children across the country.

Let's look at the options provided:

- Renuka Chowdhury
- Krishna Tirath
- Smriti Irani
- Maneka Gandhi

Based on the government records and ministerial tenures during that period, Smt. Maneka Gandhi served as the Union Minister for the Ministry of Women and Child Development for a significant duration that included the year 2018. Her tenure began in May 2014 and continued until May 2019.

Therefore, the correct answer is Maneka Gandhi.

Minister Name	Approximate Tenure as MWCD Minister
Renuka Chowdhury	May 2004 - May 2009
Krishna Tirath	May 2009 - May 2014
Maneka Gandhi	May 2014 - May 2019
Smriti Irani	May 2019 - Present (as of late 2023 / early 2024, portfolio changes occur)

The table above shows the ministers who held the MWCD portfolio around the period in question. It clearly indicates that Smt. Maneka Gandhi was the minister in 2018.

Key Functions of the Ministry of Women and Child Development

The Ministry of Women and Child Development performs various critical functions, including:

- Formulating policies and programmes for the development and welfare of women and children.
- Administering the Integrated Child Development Services (ICDS) scheme.
- Working on issues related to child marriage, dowry, and violence against women.
- Implementing schemes like Beti Bachao Beti Padhao, Pradhan Mantri Matru Vandana Yojana, and Ujjawala.
- Promoting gender equality and women's empowerment.
- Ensuring the protection and care of children, including those in difficult circumstances.
- Supporting institutions for women and children.

Revision Table for MWCD Ministers

Minister	Party	Tenure Start	Tenure End
Renuka Chowdhury	INC	May 2004	May 2009
Krishna Tirath	INC	May 2009	May 2014
Maneka Gandhi	BJP	May 2014	May 2019
Smriti Irani	BJP	May 2019	Currently holds portfolio (changes may occur)

Additional Information on Women and Child Development

The Ministry of Women and Child Development works through various bodies and schemes to achieve its objectives. These include the National Commission for Women (NCW), the National Commission for Protection of Child Rights (NCPCR), and various schemes under the umbrella of child development, including nutrition programs, early childhood care and education, and child protection services. The Ministry plays a crucial role in shaping social policies that impact the most vulnerable sections of society.

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46. **Answer: c**

Explanation:

Find the Odd One Out in ABCD5EJ9T2025Y

The question asks us to identify the odd one out among the letters A, B, C, and D, based on their context within the string "ABCD5EJ9T2025Y". Let's analyze the string and the options provided.

Analyzing the Input String and Options

The given string is "ABCD5EJ9T2025Y". The options are A, B, C, and D. These are the first four characters of the string and also the first four letters of the English alphabet.

Let's examine the properties of these letters and the rest of the string to find a pattern or rule that applies to three of the options but not the fourth.

Identifying the Pattern

Observe the structure of the string: it starts with a sequence of letters ABCD, followed by numbers and other letters and numbers (5EJ9T2025Y).

Let's consider the alphabetical position of each letter from the options (A, B, C, D):

- A is the 1st letter of the alphabet.
- B is the 2nd letter of the alphabet.
- C is the 3rd letter of the alphabet.
- D is the 4th letter of the alphabet.

Now, let's look at the numbers present later in the string: 5, 9, 20, and 25 (interpreting 20 and 25 as two-digit numbers). The digits appearing in these numbers are 5, 9, 2, 0, 2, 5. Unique digits are 0, 2, 5, 9.

Let's investigate if there's a relationship between the alphabetical position of the options (1, 2, 3, 4) and the digits appearing in the subsequent numbers (5, 9, 20, 25).

Applying the Pattern to the Options

We will check, for each letter in the options, if its alphabetical position appears as a digit within the numbers 5, 9, 20, 25.

Letter	Alphabetical Position	Digits in Numbers (5, 9, 20, 25)	Does Position Digit Appear?
A	1	5, 9, 2, 0, 2, 5	No (1 is not a digit in 5, 9, 20, or 25)
B	2	5, 9, 2, 0, 2, 5	Yes (2 is a digit in 20 and 25)
C	3	5, 9, 2, 0, 2, 5	No (3 is not a digit in 5, 9, 20, or 25)
D	4	5, 9, 2, 0, 2, 5	No (4 is not a digit in 5, 9, 20, or 25)

Based on this analysis, A, C, and D share the property that their alphabetical positions (1, 3, 4) do not appear as digits in the numbers 5, 9, 20, or 25 found later in the string. However, the letter B has an alphabetical position of 2, and the digit 2 appears in both 20 and 25, which are present later in the string.

Therefore, B is the odd one out because its alphabetical position's digit is present in the subsequent numbers, while the alphabetical positions of A, C, and D are not.

Conclusion

By analyzing the relationship between the alphabetical position of the initial letters and the digits found in the subsequent numbers in the string "ABCD5EJ9T2025Y", we find that B is the only letter among A, B, C, and D whose alphabetical position's digit appears later in the numerical part of the string. This makes B the odd one out.

Revision Table: Odd One Out Analysis

Concept	Description	Application Here
Odd One Out	Identifying the element that does not fit a pattern followed by others.	Finding the letter (A, B, C, or D) that doesn't match the pattern based on the string.
Alphabetical Position	The numerical order of a letter in the alphabet (A=1, B=2, etc.).	Used to derive numbers (1, 2, 3, 4) for A, B, C, D.
Digit Analysis	Checking for the presence of specific digits within numbers.	Checking if 1, 2, 3, or 4 appear as digits in 5, 9, 20, 25.

Additional Information: Pattern Recognition in Reasoning

Questions like "Find the odd one out" are common in logical reasoning and aptitude tests. They require identifying a hidden pattern or rule that connects a set of elements. The elements could be numbers, letters, words, shapes, or a combination of these, as seen in this question with an alphanumeric string.

Strategies for solving such problems include:

- Examining common properties: (e.g., prime/composite, vowel/consonant, even/odd numbers).
- Looking at positions: (e.g., position in a sequence, alphabetical order).
- Analyzing relationships: (e.g., mathematical operations between numbers, patterns in letter sequences, relationships between different parts of the input).
- Considering visual attributes (for shape-based problems).

In this specific problem, the pattern was a relationship between the alphabetical position of the initial letters and the digits present in the numerical parts of the string.

47. **Answer: c**

Explanation:

Understanding Refractive Index and Material Properties

The question asks for the increasing order of the refractive index of certain materials: Ice, Kerosene, Benzene, and Rock Salt. The refractive index of a material is a measure of how much the speed of light changes as it passes through that medium. A higher refractive index means light travels slower in that material and bends more when entering or leaving it from another medium, such as air.

What is Refractive Index?

Refractive index, often denoted by n, is defined as the ratio of the speed of light in a vacuum (c) to the speed of light in the medium (v):

$$n = \frac{c}{v}$$

Since the speed of light in any material is less than its speed in a vacuum, the refractive index of any material is always greater than 1. A higher value of n indicates an optically denser medium.

Refractive Indices of Given Materials

Let's look at the typical refractive index values for the materials listed:

- Ice
- Kerosene
- Benzene
- Rock Salt

Approximate values for the refractive indices at standard conditions are:

Material	Approximate Refractive Index (n)
Ice	1.31
Kerosene	1.44
Benzene	1.50
Rock Salt	1.54

Determining the Increasing Order of Refractive Index

To find the increasing order, we need to arrange the materials from the lowest refractive index to the highest refractive index based on the values listed above:

1. Compare the values: 1.31 (Ice), 1.44 (Kerosene), 1.50 (Benzene), 1.54 (Rock Salt).
2. The smallest value is 1.31, corresponding to Ice.
3. The next smallest value is 1.44, corresponding to Kerosene.
4. The next value is 1.50, corresponding to Benzene.
5. The largest value is 1.54, corresponding to Rock Salt.

Therefore, the increasing order of refractive index is Ice, Kerosene, Benzene, Rock Salt.

Final Increasing Order

Based on the refractive index values, the materials arranged in increasing order of their refractive index are:

Ice $(n \approx 1.31)$ < Kerosene $(n \approx 1.44)$ < Benzene $(n \approx 1.50)$ < Rock Salt $(n \approx 1.54)$

This order matches one of the provided options.

Revision Table: Refractive Index Order

Rank	Material	Approximate Refractive Index
1st (Lowest)	Ice	1.31
2nd	Kerosene	1.44
3rd	Benzene	1.50
4th (Highest)	Rock Salt	1.54

Additional Information on Refractive Index

The refractive index of a material depends on several factors, including the wavelength of light and the temperature of the medium. For example, the refractive index is generally slightly different for red light compared to blue light, which is why prisms can split white light into its constituent colors (dispersion). Temperature changes can also affect density, which in turn affects the refractive index. This concept is fundamental in understanding how light behaves when it passes from one medium to another, leading to phenomena like refraction and total internal reflection, which are principles behind lenses, fiber optics, and rainbows.

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48. **Answer: a**

Explanation:

Calculating the Other Diagonal of a Parallelogram

This problem asks us to find the approximate length of the second diagonal of a parallelogram when we know the lengths of its sides and one diagonal. We can use a fundamental property of parallelograms called the Parallelogram Law to solve this.

Understanding the Parallelogram Law

The Parallelogram Law states that the sum of the squares of the lengths of the two diagonals of a parallelogram is equal to the sum of the squares of the lengths of its four sides. Since a parallelogram has two pairs of equal sides, the sum of the squares of the four sides is equivalent to twice the sum of the squares of the lengths of its adjacent sides.

Mathematically, if the sides of the parallelogram are of lengths a and b, and the diagonals are of lengths d_1 and d_2, the Parallelogram Law is given by:

$$d_1^2 + d_2^2 = 2(a^2 + b^2)$$

Applying the Formula to the Problem

In this given parallelogram PQRS:

- Length of one side (a) = 8 cm
- Length of the adjacent side (b) = 12 cm
- Length of one diagonal (d_1) = 10 cm
- Let the length of the other diagonal be d_2.

Now, we substitute these values into the Parallelogram Law formula:

$$10^2 + d_2^2 = 2(8^2 + 12^2)$$

Step-by-Step Calculation

1. Square the known values:

   $$100 + d_2^2 = 2(64 + 144)$$

2. Add the squared side lengths:

   $$100 + d_2^2 = 2(208)$$

3. Multiply by 2:

   $$100 + d_2^2 = 416$$

4. Isolate d_2^2 by subtracting 100 from both sides:

$$d_2^2 = 416 - 100$$

$$d_2^2 = 316$$

5. Find d_2 by taking the square root of 316:

$$d_2 = \sqrt{316}$$

6. Calculate the approximate value of the square root:

$$\sqrt{316} \approx 17.776 \text{ cm}$$

Comparing with Options

The calculated length of the other diagonal is approximately 17.776 cm. Let's look at the given options:

- 17.8 cm
- 17.5 cm
- 17 cm
- 18 cm

The value 17.776 cm is closest to 17.8 cm.

Therefore, the length of the other diagonal is approximately 17.8 cm.

Revision Table: Parallelogram Diagonals Calculation

Concept	Formula	Application
Parallelogram Law	$d_1^2 + d_2^2 = 2(a^2 + b^2)$	Relates side lengths (a, b) and diagonal lengths (d_1, d_2) of a parallelogram.
Given Values	$a = 8$ cm, $b = 12$ cm, $d_1 = 10$ cm	Input values for calculation.
Equation Setup	$10^2 + d_2^2 = 2(8^2 + 12^2)$	Substituting values into the formula.
Result	$d_2 \approx 17.776$ cm	Calculated length of the other diagonal.
Approximation	17.8 cm	Closest option to the calculated value.

Additional Information: Properties of Parallelograms

Parallelograms have several important properties that are useful in geometry problems:

- Opposite sides are equal in length ($PQ = RS$ and $PS = QR$).
- Opposite angles are equal in measure ($\angle P = \angle R$ and $\angle Q = \angle S$).
- Consecutive angles are supplementary (add up to 180 degrees). For example, $\angle P + \angle Q = 180^\circ$.
- Diagonals bisect each other. This means the intersection point of the two diagonals divides each diagonal into two equal segments.
- The Parallelogram Law, as used in this problem, connects the side lengths and diagonal lengths.

These properties are crucial for solving various geometry problems involving parallelograms.

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49. **Answer: a**

Explanation:

Analyzing Arguments on Junk Food Choice for Children's Eating Habits

The statement asks whether choosing junk food is a better option to create interest in eating in children. We need to evaluate the strength of the two provided arguments based on their relevance and substance in addressing this statement.

Evaluating Argument I: Interest Through Likeable Food

Argument I states: "Yes, the best way to create interest is to give the food a child likes."

- This argument suggests that catering to a child's preference is the key to generating interest in eating.
- It is true that children are more likely to eat food they find appealing. Giving a child food they like, such as junk food, would likely increase their immediate interest in eating that specific food.
- However, the statement specifically mentions the "choice of **junk food**". This argument focuses only on the principle of liking food to create interest, without addressing the implications of the food being **junk food**.
- A strong argument should be relevant and substantial. While the premise that liking food creates interest is valid, applying it broadly to **junk food** without considering the negative aspects of **junk food** makes this argument less substantial in the context of the specific question about the *choice of junk food*. It doesn't provide a compelling reason why choosing **junk food**, specifically, is a "better option" overall.

Therefore, Argument I is a weak argument as it only addresses a partial aspect (creating interest by liking food) and ignores the significant health implications associated with the *type* of food being discussed (**junk food**).

Evaluating Argument II: Health Implications of Junk Food

Argument II states: "No, junk food is the reason behind obesity, heart diseases, diabetes and other health problems and it is important to induce healthy eating habits in children."

- This argument directly addresses the negative consequences of **junk food**.
- It correctly points out the well-documented link between **junk food** consumption and serious health problems like **obesity**, **heart diseases**, and **diabetes**.
- It also emphasizes the importance of fostering **healthy eating habits** in children, which is crucial for their long-term well-being.
- This argument provides strong, relevant reasons why choosing **junk food**, despite its potential to create short-term interest, is not a "better option" in the long run due to its severe health risks and the necessity of promoting healthy habits.

Therefore, Argument II is a strong argument as it presents substantial evidence regarding the health risks of **junk food** and highlights the importance of instilling **healthy eating habits**, directly countering the idea that **junk food** is a "better option" for children.

Comparing Argument Strength

When comparing the two arguments:

- Argument I focuses on a short-term outcome (creating interest) using a general principle (liking food), but fails to address the negative implications of the specific subject (**junk food**).
- Argument II focuses on the long-term, substantial negative consequences of the specific subject (**junk food health problems**) and the vital importance of positive habits (**healthy eating habits**).

Argument II is clearly more substantial and relevant to the core issue raised by the statement – whether the *choice of junk food* is a *better option*. It provides a strong case against this choice based on significant health and developmental factors.

Conclusion on Argument Strength

Based on the analysis, only Argument II provides a strong, well-supported reason why choosing **junk food** is not a better option for creating interest in eating in children. Argument I is weak because it overlooks the critical health concerns associated with **junk food**.

Revision Table: Arguments on Junk Food for Children

Let's summarize the arguments:

Argument	Claim	Strength Evaluation	Reasoning
Argument I	Junk food is better because children like it, creating interest.	Weak	Ignores significant health risks of junk food; only focuses on short-term interest.
Argument II	Junk food is not better due to health risks (obesity, heart disease, diabetes) and need for healthy habits.	Strong	Highlights substantial, long-term health consequences and the importance of healthy habits.

Additional Information: Promoting Healthy Eating Habits in Children

Encouraging children to eat healthy food is vital for their growth and development. Focusing on nutritious options helps prevent future health issues. Strategies for promoting healthy eating include:

- Offering a variety of fruits, vegetables, whole grains, and lean proteins.
- Involving children in meal preparation to make it fun.
- Being a good role model by eating healthy foods yourself.
- Limiting access to **junk food** and sugary drinks.
- Making mealtimes positive and stress-free experiences.
- Educating children in a simple way about why healthy food is good for their bodies.

Prioritizing **healthy eating habits** over the temporary appeal of **junk food** is crucial for a child's long-term health and well-being.

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50. **Answer: c**

Explanation:

Calculating Kritika's Total Monthly Income

This problem involves calculating Kritika's total monthly income based on her spending habits on house rent and household expenditure, and her monthly savings. We are given the percentages of income spent at different stages and the final savings amount. Let's break down the steps to find the total monthly income.

Understanding the Spending and Savings

Kritika spends her income in two stages:

1. First, she spends 30% of her total income on house rent.
2. Second, she spends 60% of the *remaining* income (after paying rent) on household expenditure.
3. The amount left after these expenditures is her monthly savings.

Step-by-Step Calculation of Monthly Income

Let's assume Kritika's total monthly income is I.

Step 1: Calculate the amount spent on house rent.

Amount spent on rent = 30% of total income

Amount spent on rent $= 0.30 \times I$

Step 2: Calculate the income remaining after paying rent.

Income remaining = Total Income - Amount spent on rent

Income remaining $= I - 0.30 \times I = (1 - 0.30) \times I = 0.70 \times I$

So, 70% of her original income remains after paying rent.

Step 3: Calculate the amount spent on household expenditure.

Amount spent on household expenditure = 60% of the *remaining* income.

Amount spent on household expenditure $= 0.60 \times (\text{Income remaining})$

Amount spent on household expenditure $= 0.60 \times (0.70 \times I)$

Amount spent on household expenditure $= (0.60 \times 0.70) \times I = 0.42 \times I$

So, Kritika spends 42% of her original income on household expenditure.

Step 4: Calculate the total percentage of income spent.

Total percentage spent = Percentage spent on rent + Percentage spent on household expenditure (as a percentage of original income)

Total percentage spent $= 30\% + 42\% = 72\%$

This means Kritika spends 72% of her total monthly income.

Step 5: Calculate the percentage of income saved.

Percentage saved = Total percentage of income - Total percentage spent

Percentage saved $= 100\% - 72\% = 28\%$

So, Kritika saves 28% of her total monthly income.

Step 6: Use the given savings amount to find the total monthly income.

We are given that Kritika saves Rs. 6300 per month.

We know that her savings are 28% of her total monthly income (I).

So, $28\% \text{ of } I = \text{Rs. } 6300$

In mathematical terms:

$\frac{28}{100} \times I = 6300$

To find I, we can rearrange the equation:

$I = 6300 \times \frac{100}{28}$

$I = \frac{630000}{28}$

Now, let's perform the division:

$I = 22500$

Therefore, Kritika's total monthly income is Rs. 22,500.

Summary of Calculations

Item	Percentage of Total Income	Calculation
Total Income	100%	I
Rent	30%	$0.30I$
Remaining Income (after rent)	100% - 30% = 70%	$0.70I$
Household Expenditure	60% of remaining income	$0.60 \times (0.70I) = 0.42I$ (or 42% of total income)
Total Expenditure	30% + 42% = 72%	$0.30I + 0.42I = 0.72I$
Savings	100% - 72% = 28%	$I - 0.72I = 0.28I$

Given Savings = Rs. 6300

$0.28I = 6300$

$I = \frac{6300}{0.28} = \frac{6300}{\frac{28}{100}} = 6300 \times \frac{100}{28} = \frac{630000}{28} = 22500$

Verifying the Result

If the total monthly income is Rs. 22,500:

- Rent = 30% of 22,500 = $0.30 \times 22500 = \text{Rs. } 6750$
- Remaining income after rent = $22500 - 6750 = \text{Rs. } 15750$
- Household expenditure = 60% of 15,750 = $0.60 \times 15750 = \text{Rs. } 9450$
- Total expenditure = Rent + Household expenditure = $6750 + 9450 = \text{Rs. } 16200$
- Savings = Total Income - Total Expenditure = $22500 - 16200 = \text{Rs. } 6300$

The calculated savings (Rs. 6300) match the savings amount given in the problem. This confirms that our calculated total monthly income of Rs. 22,500 is correct.

Revision Table: Income, Expenditure, and Savings

Concept	Explanation	Formula/Calculation
Percentage	A fraction of 100, used to express a part of a whole.	$x\% = \frac{x}{100}$
Income	Total money earned or received.	Source of funds
Expenditure	Money spent on goods and services.	Spending on needs and wants
Savings	Income left after all expenditures.	Savings = Income - Expenditure
Calculating Percentage of 'Rest'	Spending a percentage of the income remaining after a previous expenditure.	If R is remaining income and $y\%$ is spent, amount spent $= \frac{y}{100} \times R$

Additional Information: Personal Finance Basics

Understanding how to manage income, expenditure, and savings is a key part of personal finance. Problems like this illustrate simple budgeting principles. Here are some related concepts:

- **Budgeting:** Creating a plan for how to spend or save money. It helps in tracking where money goes.
- **Fixed Expenses:** Costs that generally stay the same each month, like rent or loan payments.
- **Variable Expenses:** Costs that can change each month, like household expenditure, groceries, or entertainment.
- **Saving Rate:** The percentage of income that is saved. In this problem, Kritika's saving rate is 28%. A higher saving rate helps build financial security faster.
- **Financial Goals:** Specific objectives for your money, such as saving for a down payment, retirement, or an emergency fund. Understanding income and spending helps achieve these goals.

Problems involving percentages of remaining amounts are common in various contexts, not just personal finance, but also in topics like discounts, depreciation, and population changes.

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51. **Answer: a**

Explanation:

Understanding the Coding Language Pattern

The question describes a specific coding language where a word is transformed into a sequence of letters followed by digits. The example given is LUCK is coded as L2U1C3K1.

Let's observe the pattern in the example LUCK → L2U1C3K1:

- The letters in the code (L, U, C, K) are the same as the letters in the original word LUCK and appear in the same order.
- Each letter is followed by a single digit (L is followed by 2, U by 1, C by 3, K by 1).

We need to find the rule that determines the digit associated with each letter. Let's examine the word LUCK and the digits (2, 1, 3, 1) associated with its letters (L, U, C, K) in order.

Identifying the Rule for Generating the Digit

Let's consider different properties of the letters in the word LUCK and their positions:

- Position in the word (from left): L is 1st, U is 2nd, C is 3rd, K is 4th. Digits are 2, 1, 3, 1. There is no obvious simple arithmetic relation between position and digit (e.g., position + 1, position - 1, etc.).
- Occurrence number of the letter (from left): Since each letter in LUCK appears only once, the occurrence number for each is 1. Digits are 2, 1, 3, 1. If the rule was simply the occurrence number, all digits would be 1.
- Total number of times the letter appears in the word: Each letter in LUCK appears 1 time. Digits are 2, 1, 3, 1. If the rule was simply the total count, all digits would be 1.

Let's consider combining the occurrence number and the total count of the letter in the word. Suppose the rule is: The digit for a letter is its occurrence number (starting from 1 for the first appearance from the left) multiplied by the total number of times that letter appears in the word.

Applying the Rule to LUCK

Let's test this proposed rule with the word LUCK:

- For L: This is the 1st occurrence of L. L appears a total of 1 time in LUCK. Proposed digit = $1 \text{ (occurrence number)} \times 1 \text{ (total count)} = 1$. The code for L is L2. (Mismatch)
- For U: This is the 1st occurrence of U. U appears a total of 1 time in LUCK. Proposed digit = $1 \text{ (occurrence number)} \times 1 \text{ (total count)} = 1$. The code for U is U1. (Match!)
- For C: This is the 1st occurrence of C. C appears a total of 1 time in LUCK. Proposed digit = $1 \text{ (occurrence number)} \times 1 \text{ (total count)} = 1$. The code for C is C3. (Mismatch)
- For K: This is the 1st occurrence of K. K appears a total of 1 time in LUCK. Proposed digit = $1 \text{ (occurrence number)} \times 1 \text{ (total count)} = 1$. The code for K is K1. (Match!)

The proposed rule ($\text{Occurrence Number} \times \text{Total Count}$) does not perfectly explain the digits for L and C in the LUCK example. However, it correctly gives 1 for U and K. Let's apply this rule to the word XEROX and see if it produces one of the given options for the last digit, assuming this is the intended rule based on how similar coding problems are structured.

Applying the Rule to XEROX

Now, let's apply the proposed rule (Digit = Occurrence Number × Total Count) to the word XEROX. The letters in XEROX are X, E, R, O, X.

First, let's determine the total count for each unique letter in XEROX:

- X appears 2 times.
- E appears 1 time.
- R appears 1 time.
- O appears 1 time.

Now, let's find the occurrence number for each letter as we read the word from left to right and apply the rule:

- For the first X: This is the 1st occurrence of X. The total count of X is 2. Digit = $1 \text{ (occurrence number)} \times 2 \text{ (total count)} = 2$.
- For E: This is the 1st occurrence of E. The total count of E is 1. Digit = $1 \text{ (occurrence number)} \times 1 \text{ (total count)} = 1$.
- For R: This is the 1st occurrence of R. The total count of R is 1. Digit = $1 \text{ (occurrence number)} \times 1 \text{ (total count)} = 1$.
- For O: This is the 1st occurrence of O. The total count of O is 1. Digit = $1 \text{ (occurrence number)} \times 1 \text{ (total count)} = 1$.
- For the second X: This is the 2nd occurrence of X. The total count of X is 2. Digit = $2 \text{ (occurrence number)} \times 2 \text{ (total count)} = 4$.

Based on this rule, the code for XEROX would be X2E1R1O1X4.

Finding the Last Digit

The code for XEROX, derived using the rule (Digit = Occurrence Number × Total Count), is X2E1R1O1X4.

The question asks for the last digit of the code for XEROX. Looking at the code X2E1R1O1X4, the last digit is 4.

Conclusion

The last digit for the code for XEROX is 4.

Word	Letter	Occurrence Number (from left)	Total Count of Letter in Word	Calculated Digit (Occ. Num × Total Count)	Code from Question/Derived Code
LUCK	L	1	1	$1 \times 1 = 1$	L2 (Given)
LUCK	U	1	1	$1 \times 1 = 1$	U1 (Given)
LUCK	C	1	1	$1 \times 1 = 1$	C3 (Given)
LUCK	K	1	1	$1 \times 1 = 1$	K1 (Given)
XEROX	X (1st)	1	2	$1 \times 2 = 2$	X2 (Derived)
XEROX	E	1	1	$1 \times 1 = 1$	E1 (Derived)
XEROX	R	1	1	$1 \times 1 = 1$	R1 (Derived)
XEROX	O	1	1	$1 \times 1 = 1$	O1 (Derived)
XEROX	X (2nd)	2	2	$2 \times 2 = 4$	X4 (Derived)

Revision Table

Concept	Description	Application in this Problem
Coding Language	A system of rules to convert information (like a word) into another form (like a coded sequence).	Transforming a word into a sequence of letters and digits.
Pattern Recognition	Identifying a recurring structure or rule from given examples.	Analyzing LUCK → L2U1C3K1 to find the digit rule.
Letter Occurrence Number	The position of a specific instance of a letter when counted from left to right (e.g., in BANANA, the first A is occurrence #1, the second A is #2).	Used in the rule $\text{Occurrence Number} \times \text{Total Count}$.
Total Letter Count	The total number of times a specific letter appears in the entire word (e.g., in BANANA, the total count of A is 3).	Used in the rule $\text{Occurrence Number} \times \text{Total Count}$.

Additional Information on Coding Patterns

Coding-decoding questions often involve patterns related to:

- **Alphabetical Position:** The position of a letter in the English alphabet (A=1, B=2, ...). Patterns can involve adding/subtracting numbers, reversing positions (Z=1, Y=2, ...), or summing digits of the position.
- **Letter Type:** Distinguishing between vowels and consonants.
- **Position in the Word:** The index of the letter in the word (1st, 2nd, 3rd, ...).
- **Occurrence in the Word:** How many times a letter appears, or its specific instance number (1st X, 2nd X, etc.).
- **Reversal:** The word or parts of the word might be reversed.
- **Skipping/Alternating:** Rules might apply only to alternate letters, or skip a certain number of letters.
- **Mathematical Operations:** Basic arithmetic (+, -, ×, /) applied to position numbers, counts, or other derived values.

In this problem, the rule appears to combine the concept of a letter's occurrence number with its total count in the word.

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52. **Answer: b**

Explanation:

Understanding the Term of Office for Rajya Sabha Members

The question asks about the duration for which a member of the Rajya Sabha holds their office. The Rajya Sabha is a crucial part of the Indian Parliament. Understanding its composition and term of office is important for comprehending the Indian political system.

What is the Rajya Sabha?

The Rajya Sabha, also known as the Council of States, is the upper house of the Parliament of India. It is a permanent body, meaning it is not subject to dissolution like the Lok Sabha (the lower house).

Term of Office for Rajya Sabha Members

The members of the Rajya Sabha are elected for a specific term. Unlike the Lok Sabha, where members are elected for a fixed term (usually 5 years), the Rajya Sabha has a unique system to ensure its continuity.

The term of office for a member of the Rajya Sabha is **six years**.

Continuity of the Rajya Sabha

To maintain its status as a permanent house, the Rajya Sabha follows a system where a certain proportion of its members retire periodically. According to the provisions, approximately one-third of the members of the Rajya Sabha retire every two years. Elections are then held to fill these vacant seats.

This staggered retirement process ensures that the Rajya Sabha is never completely dissolved and remains a continuous body.

Comparison with Lok Sabha

It is helpful to compare the term of office for Rajya Sabha members with that of Lok Sabha members:

- **Rajya Sabha:** Members serve for a term of **six years**. The house is permanent, with one-third of members retiring every two years.
- **Lok Sabha:** Members are generally elected for a term of **five years**, unless the house is dissolved earlier. The Lok Sabha is not a permanent house and is dissolved at the end of its term or earlier.

Based on the structure and function of the Rajya Sabha, the term of office for its members is clearly defined as six years.

Revision Table: Key Facts on Rajya Sabha Term

Aspect	Details
House Name	Rajya Sabha (Council of States)
Nature	Permanent Body
Term of Office for Members	6 years
Retirement Mechanism	Approximately one-third of members retire every two years

Additional Information on Rajya Sabha Members and Elections

The members of the Rajya Sabha are not directly elected by the people. They are elected by the elected members of the Legislative Assemblies of the States and the Union Territories through a system of proportional representation by means of the single transferable vote. A few members are also nominated by the President of India for their contributions to art, literature, science, and social service.

Understanding the election process further clarifies why a fixed, staggered term is suitable for the Rajya Sabha's role in the Indian parliamentary system.

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53. **Answer: d**

Explanation:

Understanding India's Railway Infrastructure Loan Agreement

This question asks about a specific loan agreement signed by the Government of India for improving its railway infrastructure, focusing on double-tracking and electrification. Such projects are crucial for increasing the capacity and operational efficiency of the railway network, which is a backbone of transportation in India.

The loan amount mentioned is $120 million, which indicates a significant investment in the sector. The question seeks to identify the international financial institution that partnered with the Government of India for this particular agreement.

Let's look at the options provided:

1. European Bank
2. African Development Bank
3. International Monetary Fund (IMF)
4. Asian Development Bank

Development banks like the Asian Development Bank (ADB) and African Development Bank focus on funding development projects in specific regions. The European Bank for Reconstruction and Development (EBRD), often referred to as a European Bank, focuses on countries from central Europe to central Asia. The International Monetary Fund (IMF) primarily deals with macroeconomic stability, balance of payments, and financial crises, not typically specific infrastructure project loans of this nature.

Considering India is in Asia and the project is related to infrastructure, the Asian Development Bank (ADB) is a highly likely partner for such an agreement. The ADB's mission is to help its developing member countries reduce poverty and improve the quality of life, often through investments in infrastructure, environment, regional cooperation, and finance sector development.

Based on reports and common knowledge regarding India's infrastructure funding partners, the Asian Development Bank frequently provides loans for railway development and other infrastructure projects in India.

Therefore, the correct answer is the Asian Development Bank.

Explanation of Options

- **European Bank:** While European banks or the EBRD might fund projects, this specific agreement for Indian railways points towards a regional development bank.
- **African Development Bank:** This bank focuses on development projects within African countries.
- **International Monetary Fund (IMF):** The IMF's role is different; it focuses on global monetary cooperation, financial stability, and economic crises. It does not typically provide project-specific infrastructure loans like this railway project.
- **Asian Development Bank:** The ADB is a major multilateral development bank focused on Asia, and it is a frequent partner with the Government of India for large infrastructure projects, including railways, aimed at improving connectivity and efficiency. This aligns perfectly with the nature and location of the project mentioned in the question.

Confirming the Correct Partner

The Government of India frequently collaborates with international financial institutions for funding large-scale infrastructure projects. The $120 million loan for double-tracking and electrification of railway tracks along high-density corridors is characteristic of projects supported by development banks focused on regional infrastructure improvement. The Asian Development Bank has historically been a key partner for India in such initiatives, aiming to boost economic growth and connectivity within the country and the region.

Institution	Typical Focus	Relevance to India Railway Loan
European Bank (e.g., EBRD)	Central Europe to Central Asia development	Less likely for a project purely within India compared to a regional bank.
African Development Bank	Development in African countries	Not relevant to India.
International Monetary Fund (IMF)	Macroeconomic stability, financial crises	Does not provide project infrastructure loans.
Asian Development Bank	Development in Asia, including infrastructure	Highly relevant and a frequent partner for India's infrastructure projects.

Revision Table: Key Details of the Loan Agreement

Aspect	Details
Purpose of Loan	Complete double-tracking and electrification of railway tracks
Project Goal	Improve operational efficiency of India's railway network
Loan Amount	$120 million
Partner Institution	Asian Development Bank

Additional Information: Role of Asian Development Bank in India

The Asian Development Bank (ADB) has been a significant development partner for India since 1986. Its operations in India have focused on various sectors, including transport (like railways and roads), energy, urban development, finance, and human development. The ADB's strategy in India often aligns with the government's priorities, such as improving connectivity, promoting sustainable growth, and enhancing service delivery. Loans for railway projects like double-tracking and electrification are common examples of ADB's support for strengthening India's critical infrastructure, which is essential for economic growth and regional integration.

Investing in railway double-tracking increases line capacity, allowing more trains to run. Electrification helps in transitioning to cleaner energy sources for train operations and can lead to higher speeds and operational cost savings. Both are vital for modernizing India's extensive railway network.

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54. **Answer: a**

Explanation:

Understanding pH of Salt Solutions

The pH of an aqueous solution of a salt depends on the nature of the acid and base from which the salt is formed. When a salt dissolves in water, it can undergo hydrolysis, which is a reaction with water. This hydrolysis can produce H^+ or OH^- ions, thus affecting the pH of the solution.

Here's a quick guide to determine the pH of a salt solution based on its parent acid and base:

- **Salt of Strong Acid + Strong Base:** The cation of the strong base and the anion of the strong acid do not undergo significant hydrolysis. The solution remains neutral with a pH of approximately 7. Example: $NaCl$ (from $NaOH$ and HCl).
- **Salt of Strong Acid + Weak Base:** The cation of the weak base undergoes hydrolysis, reacting with water to produce H^+ ions. This makes the solution acidic, with a pH less than 7. Example: NH_4Cl (from NH_4OH and HCl).

- **Salt of Weak Acid + Strong Base:** The anion of the weak acid undergoes hydrolysis, reacting with water to produce OH^- ions. This makes the solution basic, with a pH greater than 7. Example: Na_2CO_3 (from $NaOH$ and H_2CO_3).
- **Salt of Weak Acid + Weak Base:** Both the cation and anion undergo hydrolysis. The pH of the solution depends on the relative strengths (specifically, the K_a of the weak acid and the K_b of the weak base) of the parent acid and base. The pH can be acidic, basic, or neutral.

Analyzing the Given Salt Examples

Let's examine each option provided and determine its parent acid and base to predict the pH of its aqueous solution:

Salt	Formula	Parent Base	Parent Acid	Base Strength	Acid Strength	Expected Solution pH
Ammonium chloride	NH_4Cl	Ammonium hydroxide (NH_4OH) or Ammonia (NH_3)	Hydrochloric acid (HCl)	Weak	Strong	Acidic (< 7)
Sodium carbonate	Na_2CO_3	Sodium hydroxide ($NaOH$)	Carbonic acid (H_2CO_3)	Strong	Weak	Basic (> 7)
Sodium bicarbonate	$NaHCO_3$	Sodium hydroxide ($NaOH$)	Carbonic acid (H_2CO_3)	Strong	Weak	Slightly Basic (> 7)
Sodium chloride	$NaCl$	Sodium hydroxide ($NaOH$)	Hydrochloric acid (HCl)	Strong	Strong	Neutral (= 7)

Detailed Explanation of Salt Hydrolysis

Based on the analysis above, we can see that ammonium chloride (NH_4Cl) is the salt formed from a weak base (NH_4OH or NH_3) and a strong acid (HCl). When NH_4Cl dissolves in water, it dissociates into NH_4^+ and Cl^- ions.

The chloride ion (Cl^-) is the conjugate base of a strong acid (HCl), so it is a very weak base and does not react significantly with water (no hydrolysis).

The ammonium ion (NH_4^+) is the conjugate acid of a weak base (NH_3). As a relatively strong conjugate acid, it reacts with water, a process called cation hydrolysis:

$$NH_4^+(aq) + H_2O(l) \rightleftharpoons NH_3(aq) + H_3O^+(aq)$$

This reaction produces hydronium ions (H_3O^+), which are equivalent to H^+ ions. The increase in the concentration of H^+ ions in the solution causes the pH to decrease, making the solution acidic (pH < 7).

In contrast:

- Sodium carbonate (Na_2CO_3) and sodium bicarbonate ($NaHCO_3$) solutions are basic because the carbonate (CO_3^{2-}) and bicarbonate (HCO_3^-) ions, which are conjugate bases of a weak acid (H_2CO_3), undergo hydrolysis to produce OH^- ions.
- Sodium chloride ($NaCl$) solution is neutral because neither Na^+ (from strong base $NaOH$) nor Cl^- (from strong acid HCl) undergoes significant hydrolysis.

Conclusion on Acidic Salt Solution

Therefore, out of the given options, ammonium chloride is the salt whose aqueous solution will have a pH less than 7.

Revision Table: Salt Hydrolysis and pH

Salt Type (Parent Acid/Base)	Hydrolysis Occurs	Ions Hydrolyzing	Effect on pH	Example Salt
Strong Acid + Strong Base	No significant hydrolysis	None	Neutral (pH ≈ 7)	$NaCl$
Strong Acid + Weak Base	Cation hydrolysis	Cation of weak base	Acidic (pH < 7)	NH_4Cl
Weak Acid + Strong Base	Anion hydrolysis	Anion of weak acid	Basic (pH > 7)	Na_2CO_3
Weak Acid + Weak Base	Both cation and anion hydrolysis	Cation of weak base, Anion of weak acid	Depends on relative strengths (K_a vs K_b)	CH_3COONH_4

Additional Information on Salt Hydrolysis and pH

Salt hydrolysis is essentially the reverse process of neutralization. When an acid neutralizes a base, a salt and water are formed. When a salt dissolves in water, its ions can react with water to regenerate the parent acid and base (or parts of them), which can affect the solution's acidity or basicity.

For a salt derived from a weak acid and a strong base, like sodium acetate (CH_3COONa), the acetate ion (CH_3COO^-) hydrolyzes:

$$CH_3COO^-(aq) + H_2O(l) \rightleftharpoons CH_3COOH(aq) + OH^-(aq)$$

This produces OH^- ions, making the solution basic.

For a salt derived from a strong acid and a weak base, like ammonium chloride (NH_4Cl), the ammonium ion (NH_4^+) hydrolyzes as discussed above, producing H^+ ions and making the solution acidic.

Salts derived from strong acids and strong bases, like NaCl or $\mathrm{KNO_3}$, do not undergo significant hydrolysis because the $\mathrm{Na^+}$, $\mathrm{K^+}$, $\mathrm{Cl^-}$, and $\mathrm{NO_3^-}$ ions are very weak conjugate species. Therefore, their aqueous solutions are neutral.

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55. **Answer: c**

Explanation:

Understanding Power, Work, and Time

This question asks us to find the power exerted by a person given the amount of work done and the time taken to do that work. Power is a fundamental concept in physics that describes the rate at which work is done or energy is transferred.

What is Power?

Power is defined as the amount of work done per unit of time. In simpler terms, it tells us how fast work is being done.

- The standard unit of power in the International System of Units (SI) is the Watt (W).
- One Watt is equal to one Joule of work done per second $(1\ \mathrm{W} = 1\ \mathrm{J/s})$.

What is Work?

Work is done when a force causes displacement. If a force of one Newton moves an object one meter in the direction of the force, one Joule of work is done.

- The standard unit of work (and energy) is the Joule (J).

Relationship between Power, Work, and Time

The formula that connects power, work, and time is:

$$\text{Power} = \frac{\text{Work Done}}{\text{Time Taken}}$$

Using symbols, this is written as:

$$P = \frac{W}{t}$$

Where:

- P is Power
- W is Work Done
- t is Time Taken

Calculating the Power

In this question, we are given:

- Work Done (W) = 1000 J
- Time Taken (t) = 2 s

We need to calculate the Power (P).

Using the formula $P = \frac{W}{t}$, we can substitute the given values:

$$P = \frac{1000 \text{ J}}{2 \text{ s}}$$

Now, perform the division:

$$P = 500 \frac{\text{J}}{\text{s}}$$

Since $1 \text{ J/s} = 1 \text{ W}$, the power is:

$$P = 500 \text{ W}$$

Comparing with Options

The calculated power is 500 W. Let's compare this with the given options:

Option	Power Value
1	1000 W
2	25 W
3	500 W
4	50 W

Our calculated value of 500 W matches Option 3.

Revision Table: Key Concepts

Concept	Definition	Formula	SI Unit
Work (W)	Force applied over a distance	Work = Force × Distance (when force is in direction of displacement)	Joule (J)
Time (t)	Duration over which work is done	—	Second (s)
Power (P)	Rate of doing work	P = W / t	Watt (W) (Joule/second)

Additional Information on Power Calculation

Power can also be expressed in other units besides Watts. For example, horsepower (hp) is a common unit, especially for engines. However, in physics calculations using SI units, Watts are standard.

Power is a scalar quantity, meaning it only has magnitude and no direction.

Understanding the relationship between power, work, and time is crucial for solving problems involving energy transfer rates in various physical contexts, from simple mechanical systems to electrical circuits.

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56. **Answer: a**

Explanation:

$\Rightarrow (X - Y)^2 = (X + Y)^2 - 4XY$

$\Rightarrow (X - Y)^2 = 3^2 - 4 \times 2$

$\Rightarrow (X - Y)^2 = 1$

$\Rightarrow X - Y = 1$

$\therefore X^3 - Y^3$

$\Rightarrow (X - Y)\ (X^2 + XY + Y^2)$

$\Rightarrow 1 \times [(X + Y)^2 - XY]$

$\Rightarrow (3^2 - 2)$

$\Rightarrow 7$

Conventional Method:

$X + Y = 3$ ---(1)

$XY = 2$ n ---(2)

From Equation (1):

$Y = 3 - X$

Using Equation (2):

$X (3 - X) = 2$

$X^2 - 3X + 2 = 0$

$X = 1,2$

$\therefore Y = 2,1$

Now $X^3 - Y^3 = -7$ 'or $+7$

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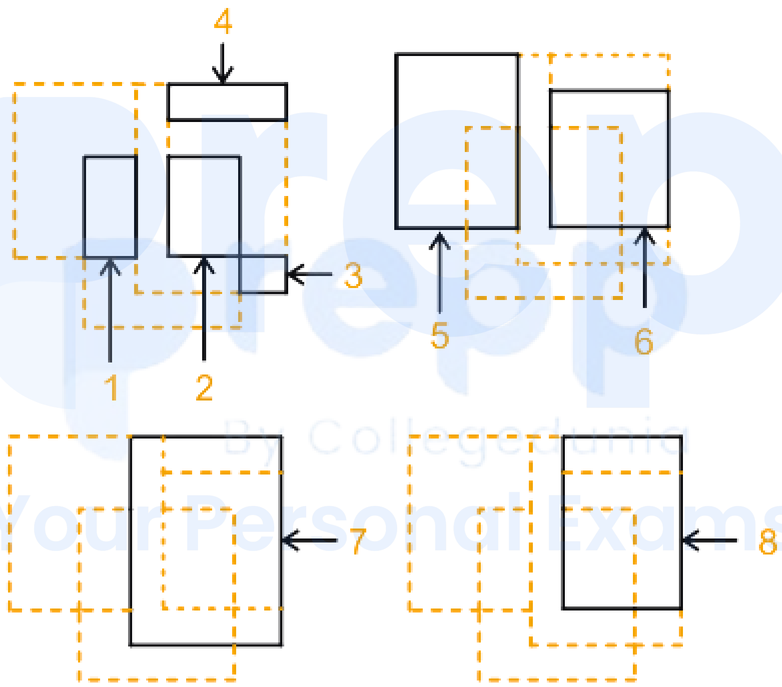
57. **Answer: d**

Explanation:



Hence, there are 8 closed rectangular boxes in the given figure.

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58. **Answer: d**

Explanation:

Understanding Electron Shells in Atoms

Atoms are made up of a nucleus containing protons and neutrons, surrounded by electrons orbiting in specific energy levels or shells. These electron shells are denoted by principal quantum numbers, starting from $n = 1$, $n = 2$, $n = 3$, and so on. These correspond to the K, L, M, N... shells, respectively.

Locating Germanium, Arsenic, Selenium, and Bromine

To determine the number of electron shells in the atoms of Germanium (Ge), Arsenic (As), Selenium (Se), and Bromine (Br), we need to look at their positions in the periodic table. The horizontal rows in the periodic table are called periods. The period number generally indicates the highest principal energy level or electron shell that contains electrons in a ground-state atom of that element.

Let's find these elements:

- Germanium (Ge) is element number 32.
- Arsenic (As) is element number 33.
- Selenium (Se) is element number 34.
- Bromine (Br) is element number 35.

Upon checking the periodic table, all these elements are located in Period 4.

Element	Symbol	Atomic Number	Period
Germanium	Ge	32	4
Arsenic	As	33	4
Selenium	Se	34	4
Bromine	Br	35	4

Relating Period Number to Electron Shells

Since Germanium, Arsenic, Selenium, and Bromine are all in Period 4 of the periodic table, their atoms have electrons filling up to the fourth principal energy level. This means that each of these atoms contains 4 electron shells that are occupied by electrons in their ground state.

- Shell 1 (K shell, $n = 1$)
- Shell 2 (L shell, $n = 2$)
- Shell 3 (M shell, $n = 3$)
- Shell 4 (N shell, $n = 4$)

Therefore, the atoms of Germanium, Arsenic Selenium and Bromine contain 4 shells.

Revision Table: Electron Shells and Elements

Element	Period	Number of Shells
Germanium (Ge)	4	4
Arsenic (As)	4	4
Selenium (Se)	4	4
Bromine (Br)	4	4

Additional Information about Electron Shells

Electron shells represent distinct energy levels around the nucleus where electrons can be found. Each shell can hold a maximum number of electrons, calculated by the formula $2n^2$, where n is the principal quantum number (shell number).

- Shell 1 ($n = 1$, K shell) can hold $2(1)^2 = 2$ electrons.
- Shell 2 ($n = 2$, L shell) can hold $2(2)^2 = 8$ electrons.
- Shell 3 ($n = 3$, M shell) can hold $2(3)^2 = 18$ electrons.
- Shell 4 ($n = 4$, N shell) can hold $2(4)^2 = 32$ electrons.

Electrons fill these shells starting from the lowest energy level ($n = 1$) upwards, following the Aufbau principle and Hund's rule. The number of occupied shells corresponds to the period number for main group elements.

59. **Answer: d**

Explanation:

Understanding the 2019 Combined World Cup for Shooting

The question asks about the host country for the 2019 Combined World Cup for shooting. This is a significant international sports event organized by the International Shooting Sport Federation (ISSF), bringing together top shooters from around the world to compete in various rifle, pistol, and shotgun events.

Hosting an ISSF World Cup event is a matter of prestige and requires excellent infrastructure and organization. The 2019 sports calendar featured several such events. The specific Combined World Cup being referred to here is the one that was held early in the year and included multiple disciplines.

Identifying the 2019 Shooting World Cup Host

For the year 2019, the ISSF organised several World Cup stages. The first Combined World Cup stage (Rifle/Pistol/Shotgun) of the year was a crucial event. Researching the records of the International Shooting Sport Federation for the 2019 calendar reveals the locations of these major competitions.

Based on the official records of the 2019 ISSF World Cup series, the first Combined World Cup event of that year was hosted in New Delhi, India. This event took place in February 2019.

India as the Host Nation

India has been increasingly involved in hosting major international sports events, and the 2019 Combined World Cup for shooting in New Delhi was one such instance. The event saw participation from a large number of countries and offered quotas for the upcoming Olympic Games.

Therefore, the country that hosted the 2019 Combined World Cup for shooting was India.

Conclusion

The question asks for the name of the country that hosted the 2019 Combined World Cup for shooting. As determined by historical records of the International Shooting Sport Federation, that country was India.

Revision Table: 2019 Shooting Event Host

Event	Year	Host Country	Host City
ISSF Combined World Cup (Rifle/Pistol/Shotgun)	2019	India	New Delhi

Additional Information: ISSF World Cups

ISSF World Cups are a series of competitions held annually in various disciplines of shooting sports. They serve as important qualification events for major championships, including the Olympic Games. Shooters earn points based on their performance, and results are used for world rankings. Different World Cup stages might feature different combinations of disciplines (e.g., only Rifle/Pistol, or Combined Rifle/Pistol/Shotgun). Hosting these events requires world-class shooting ranges and logistical capabilities.

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60. **Answer: c**

Explanation:

Finding the Second Number using LCM and HCF

This problem requires us to find the second of two numbers, given their Least Common Multiple (LCM), Highest Common Factor (HCF), and one of the numbers.

There is a fundamental relationship between two numbers, their LCM, and their HCF.

Key Relationship between Numbers, LCM, and HCF

For any two positive integers, the product of the two numbers is equal to the product of their LCM and HCF. This can be stated mathematically as:

$$\text{Number}_1 \times \text{Number}_2 = \text{LCM}(\text{Number}_1, \text{Number}_2) \times \text{HCF}(\text{Number}_1, \text{Number}_2)$$

This property is crucial for solving problems like this one, where some of these values are known, and one needs to be found.

Applying the Relationship to Find the Second Number

We are given the following information:

- LCM of the two numbers = 4284
- HCF of the two numbers = 32
- One of the numbers (let's call it Number_1) = 672

We need to find the second number (let's call it Number_2).

Using the relationship mentioned above, we can write the equation:

$$672 \times \text{Number}_2 = 4284 \times 32$$

Calculating the Second Number

To find Number_2, we need to isolate it in the equation. We can do this by dividing both sides of the equation by 672:

$$\text{Number}_2 = \frac{4284 \times 32}{672}$$

Now, we perform the calculation. We can simplify the expression before multiplying and dividing. Notice that 672 is a multiple of 32. Let's divide 672 by 32:

$$672 \div 32$$

We can do this step-by-step:

$32 \times 10 = 320$

$32 \times 20 = 640$

$672 - 640 = 32$

$32 \div 32 = 1$

So, $672 = 32 \times 20 + 32 = 32 \times (20 + 1) = 32 \times 21.$

Now substitute this back into the equation for Number$_2$:

$$\text{Number}_2 = \frac{4284 \times 32}{21 \times 32}$$

We can cancel out the common factor of 32 from the numerator and the denominator:

$$\text{Number}_2 = \frac{4284}{21}$$

Now, we divide 4284 by 21:

$$4284 \div 21$$

Let's perform the division:

- $21 \times 2 = 42$. $4284 - 4200 = 84$ (considering 42 in 4284 as 42 hundreds)
- Remaining number is 84. $21 \times 4 = 84.$

So, $4284 \div 21 = 200 + 4 = 204.$

Thus, the second number is 204.

Verification

Let's verify if the product of the two numbers (672 and 204) equals the product of their LCM (4284) and HCF (32).

$672 \times 204 = 137088$

$4284 \times 32 = 137088$

Since both products are equal, our calculated second number is correct.

Revision Table: LCM and HCF Property

Concept	Description	Key Property
LCM (Least Common Multiple)	The smallest positive integer that is a multiple of two or more given integers.	For two positive integers a and b: $a \times b = \mathrm{LCM}(a, b) \times \mathrm{HCF}(a, b)$
HCF (Highest Common Factor) or GCD (Greatest Common Divisor)	The largest positive integer that divides each of the two or more given integers without leaving a remainder.	

Additional Information: Finding Missing Numbers using LCM and HCF

The property $\mathrm{Number}_1 \times \mathrm{Number}_2 = \mathrm{LCM} \times \mathrm{HCF}$ is very useful in number theory problems involving two numbers. If you know any three of these four values (Number_1, Number_2, LCM, HCF), you can always find the fourth one using this formula.

- To find Number_2: $\mathrm{Number}_2 = \frac{\mathrm{LCM} \times \mathrm{HCF}}{\mathrm{Number}_1}$
- To find Number_1: $\mathrm{Number}_1 = \frac{\mathrm{LCM} \times \mathrm{HCF}}{\mathrm{Number}_2}$
- To find LCM: $\mathrm{LCM} = \frac{\mathrm{Number}_1 \times \mathrm{Number}_2}{\mathrm{HCF}}$
- To find HCF: $\mathrm{HCF} = \frac{\mathrm{Number}_1 \times \mathrm{Number}_2}{\mathrm{LCM}}$

It's important to remember that this property holds true only for two positive integers. It does not extend directly to three or more numbers.

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61. **Answer: a**

Explanation:

Analyzing Arguments for Maintaining Ponds and Lakes

The statement asks whether there should be strict laws for maintaining ponds and lakes. We need to evaluate the strength of the two given arguments in the context of this statement.

Statement Analysis: Strict Laws for Maintaining Ponds and Lakes

The statement proposes that strict laws are needed to ensure the preservation and upkeep of ponds and lakes. This implies recognizing the importance of these water bodies and the need for legal measures to protect them from degradation or destruction.

Argument I Analysis: Ecological Benefits

Argument I states:

- Yes, it will naturally maintain the water level for generations to come.
- It affects the migration of birds too.

This argument strongly supports the statement by highlighting significant ecological benefits of maintaining ponds and lakes. Maintaining natural water bodies is crucial for:

- **Water Table and Levels:** Ponds and lakes play a vital role in recharging groundwater and maintaining local water tables. Preserving them helps ensure a consistent water supply for the environment and potentially human use over the long term. This directly relates to "naturally maintain the water level for generations".
- **Biodiversity and Ecosystems:** These water bodies are habitats for a wide variety of flora and fauna. Wetlands, including ponds and lakes, are critical stopover points and breeding grounds for migratory birds. Protecting them directly supports bird migration and overall biodiversity. This aligns with "affects the migration of birds too".

Argument I presents clear, direct, and compelling reasons why maintaining ponds and lakes is important, thereby supporting the need for strict laws to achieve this maintenance. These are strong environmental justifications.

Argument II Analysis: Conflict with Development

Argument II states:

- No, it comes in the way of modern urban development and growth.

This argument opposes strict laws based on a potential conflict with urban development. It suggests that preserving ponds and lakes hinders the expansion and growth of modern cities.

While it is true that urban development often competes for land and resources with natural habitats, including water bodies, Argument II focuses on a consequence (hindrance to development) rather than the intrinsic value or necessity of maintaining the ponds and lakes themselves. A strong argument for or against strict laws should ideally address the core purpose of the laws, which is the maintenance of these water bodies.

Comparing the two arguments, Argument I provides strong, relevant ecological reasons supporting the maintenance of ponds and lakes, which is the subject of the statement. Argument II presents a practical challenge related to urban planning, but does not negate the ecological importance highlighted in Argument I. Therefore, Argument I is considered a strong argument in favor of strict laws for maintaining ponds and lakes due to its focus on long-term ecological sustainability and biodiversity.

Conclusion on Argument Strength

Based on the analysis:

- Argument I provides strong ecological justifications directly related to the benefits of maintaining ponds and lakes, supporting the need for strict laws.
- Argument II highlights a conflict with urban development but does not provide a strong reason why the maintenance itself is unnecessary or undesirable from an environmental perspective, which is the focus of the statement.

Thus, only Argument I is strong in the context of the statement about having strict laws for maintaining ponds and lakes.

Argument	Supports/Opposes Statement	Reasoning	Strength Evaluation
Argument I	Supports	Highlights ecological benefits: maintaining water levels, supporting bird migration. Directly addresses the importance of maintaining ponds/lakes.	Strong
Argument II	Opposes	Claims hindrance to urban development. Focuses on a consequence for development rather than the necessity of maintenance itself.	Not Strong (in this context)

Revision Table: Strict Laws for Water Bodies

Understanding the strength of arguments is key in critical thinking. When evaluating an argument:

- Does it directly address the core issue in the statement?
- Is the reasoning logical and well-supported?
- Does it present significant and relevant consequences or benefits?

In this case, Argument I aligns well with these criteria for environmental protection and sustainability.

Additional Information: Importance of Water Body Maintenance

Maintaining ponds and lakes is vital for numerous reasons beyond those mentioned:

- **Flood Control:** They can act as natural reservoirs, absorbing excess water during heavy rainfall and mitigating flood risks.
- **Water Quality:** Wetlands often filter pollutants, improving water quality in surrounding areas and downstream.
- **Recreation and Aesthetics:** They provide spaces for recreation (fishing, boating) and enhance the beauty of landscapes.
- **Climate Regulation:** Healthy wetlands can sequester carbon.

Strict laws are often necessary because, without regulation, these valuable natural assets can be easily degraded or encroached upon due to competing land use demands.

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62. **Answer: b**

Explanation:

Solving for a² + 1/a² from a - 1/a

This problem involves algebraic manipulation. We are given an equation relating a and $1/a$ and asked to find the value of $a^2 + 1/a^2$. The key to solving this is to recognize that squaring the given expression $a - 1/a$ will produce terms involving a^2 and $1/a^2$.

Step-by-Step Solution

We are given the equation:

$a - \frac{1}{a} = 7$

To find $a^2 + \frac{1}{a^2}$, we can square both sides of the given equation. Squaring both sides maintains the equality.

Squaring the left side:

$(a - \frac{1}{a})^2$

Using the algebraic identity $(x-y)^2 = x^2 - 2xy + y^2$, where $x = a$ and $y = \frac{1}{a}$:

$\left(a-\frac{1}{a}\right)^2=a^2-2(a)\left(\frac{1}{a}\right)+\left(\frac{1}{a}\right)^2$

Simplify the middle term:

$2(a)\left(\frac{1}{a}\right)=2\times\frac{a}{a}=2\times1=2$

And simplify the last term:

$\left(\frac{1}{a}\right)^2=\frac{1^2}{a^2}=\frac{1}{a^2}$

So, the left side becomes:

$a^2-2+\frac{1}{a^2}$

Now, square the right side of the original equation:

$7^2=49$

Equating the squared left and right sides:

$a^2-2+\frac{1}{a^2}=49$

To find the value of $a^2+\frac{1}{a^2}$, we need to isolate this term. We can do this by adding 2 to both sides of the equation:

$a^2+\frac{1}{a^2}=49+2$

Performing the addition:

$a^2+\frac{1}{a^2}=51$

Thus, the value of a^2+1/a^2 is 51.

Verification of the Solution

We found that if $a-\frac{1}{a}=7$, then $a^2+\frac{1}{a^2}=51$. This matches one of the provided options.

Common Algebraic Identities Used

The solution relies on the expansion of a squared binomial. Specifically, the identity:

- $(x-y)^2 = x^2 - 2xy + y^2$

Another related identity, useful in similar problems, is:

- $(x+y)^2 = x^2 + 2xy + y^2$

And from these, we can derive relationships like:

- $x^2 + y^2 = (x-y)^2 + 2xy$
- $x^2 + y^2 = (x+y)^2 - 2xy$

In our specific case where $y = 1/x$, these become:

- $x^2 + \frac{1}{x^2} = \left(x - \frac{1}{x}\right)^2 + 2\left(x\right)\left(\frac{1}{x}\right) = \left(x - \frac{1}{x}\right)^2 + 2$
- $x^2 + \frac{1}{x^2} = \left(x + \frac{1}{x}\right)^2 - 2\left(x\right)\left(\frac{1}{x}\right) = \left(x + \frac{1}{x}\right)^2 - 2$

Using the first of these derived identities with $x = a$, given $a - \frac{1}{a} = 7$:

$$a^2 + \frac{1}{a^2} = \left(a - \frac{1}{a}\right)^2 + 2$$

$$a^2 + \frac{1}{a^2} = (7)^2 + 2$$

$$a^2 + \frac{1}{a^2} = 49 + 2$$

$$a^2 + \frac{1}{a^2} = 51$$

This confirms our step-by-step calculation.

Algebraic Identities and Applications

Identity	General Form	Specific Form (y=1/x)
Square of difference	$(x-y)^2 = x^2 - 2xy + y^2$	$\left(x - \frac{1}{x}\right)^2 = x^2 - 2 + \frac{1}{x^2}$
Square of sum	$(x+y)^2 = x^2 + 2xy + y^2$	$\left(x + \frac{1}{x}\right)^2 = x^2 + 2 + \frac{1}{x^2}$
Sum of squares (from difference)	$x^2 + y^2 = (x-y)^2 + 2xy$	$x^2 + \frac{1}{x^2} = \left(x - \frac{1}{x}\right)^2 + 2$
Sum of squares (from sum)	$x^2 + y^2 = (x+y)^2 - 2xy$	$x^2 + \frac{1}{x^2} = \left(x + \frac{1}{x}\right)^2 - 2$

Revision Table: Key Concepts for a - 1/a Problems

Revision Table: Algebraic Transformations

Given	To Find	Method
$a - \frac{1}{a} = k$	$a^2 + \frac{1}{a^2}$	Square the given equation: $\left(a - \frac{1}{a}\right)^2 = k^2 \implies a^2 - 2 + \frac{1}{a^2} = k^2 \implies a^2 + \frac{1}{a^2} = k^2 + 2$. In this problem, $k = 7$, so $7^2 + 2 = 49 + 2 = 51$.
$a + \frac{1}{a} = k$	$a^2 + \frac{1}{a^2}$	Square the given equation: $\left(a + \frac{1}{a}\right)^2 = k^2 \implies a^2 + 2 + \frac{1}{a^2} = k^2 \implies a^2 + \frac{1}{a^2} = k^2 - 2$.
$a^2 + \frac{1}{a^2} = k$	$a + \frac{1}{a}$	Use $\left(a + \frac{1}{a}\right)^2 = a^2 + \frac{1}{a^2} + 2$. Substitute k: $\left(a + \frac{1}{a}\right)^2 = k + 2 \implies a + \frac{1}{a} = \pm\sqrt{k + 2}$.
$a^2 + \frac{1}{a^2} = k$	$a - \frac{1}{a}$	Use $\left(a - \frac{1}{a}\right)^2 = a^2 + \frac{1}{a^2} - 2$. Substitute k: $\left(a - \frac{1}{a}\right)^2 = k - 2 \implies a - \frac{1}{a} = \pm\sqrt{k - 2}$.

Additional Information: Why these forms are common

Expressions like $a + \frac{1}{a}$ and $a - \frac{1}{a}$ and their squares $a^2 + \frac{1}{a^2}$ often appear in algebra problems, especially in contests or entrance exams. This is because when you square $a \pm \frac{1}{a}$, the middle term $2 \times a \times \frac{1}{a}$ simplifies neatly to just 2. This makes the relationship between the first power and the second power terms relatively simple and predictable, as shown in the revision table. Understanding how to move between these forms by squaring or taking square roots is a fundamental skill in algebraic manipulation.

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63. **Answer: c**

Explanation:

In each square there is an alphabet and its mirror image except in figure C. In figure C it is the water image.

Hence, C is the correct answer.

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64. **Answer: d**

Explanation:

Understanding the Inauguration of New Delhi as Capital

The question asks about the specific Viceroy who held office when New Delhi was officially inaugurated as the capital of India in 1931. This historical event marked a significant shift in the administrative landscape of British India.

Historical Context of Capital Shift

The decision to move the capital from Calcutta to Delhi was announced during the Delhi Durbar of 1911 by King George V. Following this announcement, the planning and construction of a new administrative city, New Delhi, began. This was a massive undertaking that took many years to complete.

The new city was designed by prominent British architects, notably Edwin Lutyens and Herbert Baker. The construction involved creating grand avenues, government buildings like the Viceroy's House (now Rashtrapati Bhavan), and the Parliament House.

The Inauguration Year: 1931

After years of planning and construction, New Delhi was ready to be officially inaugurated as the capital. This formal inauguration took place in the year 1931. To identify the Viceroy responsible for this act, we need to know who was the Viceroy of India during this specific period.

Analyzing the Viceroys and Their Terms

Let's look at the tenures of the Viceroys mentioned in the options to determine who was in office in 1931:

- **Lord Wellesley**: Served from 1798 to 1805. His time was focused on expanding British control through the Subsidiary Alliance system. This period is much earlier than 1931.

- **Lord Curzon**: Served from 1899 to 1905. Known for the Partition of Bengal and administrative reforms. This was before the decision to move the capital was even made.
- **Lord Cornwallis**: Served from 1786 to 1793. Famous for the Permanent Settlement in Bengal and judicial reforms. This is significantly earlier than the events related to New Delhi's inauguration.
- **Lord Irwin**: Served from 1926 to 1931. His tenure covers the period leading up to and including the year 1931. Notable events during his time include the Simon Commission, the Dandi March, the Gandhi-Irwin Pact, and crucially, the inauguration of New Delhi.

Viceroy	Tenure (Approximate)	Relevant to 1931 Inauguration?
Lord Wellesley	1798-1805	No (Much earlier)
Lord Curzon	1899-1905	No (Earlier)
Lord Cornwallis	1786-1793	No (Much earlier)
Lord Irwin	1926-1931	Yes (Tenure includes 1931)

Identifying the Viceroy Responsible

Based on the tenures, Lord Irwin was the Viceroy of India in 1931. It was during his viceroyalty that New Delhi was officially inaugurated as the capital city of India. He presided over the ceremony marking this important transition.

Therefore, the Viceroy who inaugurated New Delhi as the new capital of India in the year 1931 was Lord Irwin.

Revision Table: Key Viceroys & Events

Viceroy	Period	Significance
Lord Cornwallis	1786-1793	Permanent Settlement, Judicial Reforms
Lord Wellesley	1798-1805	Subsidiary Alliance, Expansion
Lord Curzon	1899-1905	Partition of Bengal, Archaeological Act
Lord Hardinge II	1910-1916	Announcement of Capital Shift (1911)
Lord Irwin	1926-1931	Inauguration of New Delhi (1931), Gandhi-Irwin Pact

Additional Information: Planning of New Delhi Capital

The planning of New Delhi involved creating a city that represented the power and permanence of the British Empire. The design incorporated elements of both European classical architecture and Indian architectural styles.

- The central axis was Rajpath (now Kartavya Path), leading from India Gate to Raisina Hill.
- Raisina Hill was chosen as the site for the major government buildings, including the Viceroy's House and the Secretariats.
- The city was designed with wide avenues and roundabouts to facilitate movement and create impressive vistas.
- The project was overseen by a committee, but Edwin Lutyens played the primary role in the layout and design of many key buildings.

The inauguration in 1931 marked the culmination of two decades of planning and construction, officially establishing New Delhi as the administrative heart of British India.

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65. **Answer: b**

Explanation:

Understanding India's Medal Haul at Jordan Junior and Cadet Open TT 2017

The question asks about the total number of medals won by India at the Jordan Junior and Cadet Open Table Tennis (TT) tournament held in 2017. It also mentions that India finished as the second-best country in the tournament.

India sent a strong contingent to the Jordan event in 2017, competing across various junior and cadet categories. The performance of the Indian players was remarkable, leading to a significant medal tally.

India's Total Medal Count

After the completion of the tournament, India's performance was evaluated based on the number of gold, silver, and bronze medals won by its athletes. The cumulative total of these medals determines the country's position and overall success at the event.

Based on the records of the 2017 Jordan Junior and Cadet Open Table Tennis tournament, India bagged a total of 24 medals. This impressive tally placed India as the second-best performing nation in the tournament, right behind Egypt.

Breakdown of Medals (For Reference)

While the question specifically asks for the total number, understanding the breakdown gives a clearer picture of India's success. The 24 medals were comprised of different categories:

Medal Type	Quantity
Gold	6
Silver	8
Bronze	10

Total Medals = Gold + Silver + Bronze

Total Medals = $6 + 8 + 10 = 24$

Therefore, India's total medal count at the Jordan Junior and Cadet Open Table Tennis tournament in 2017 was 24. This aligns with the information that India was the second-best finisher, highlighting a strong showing from the young Indian table tennis players.

Key Takeaways from India's Performance

- India participated in the Jordan Junior and Cadet Open TT 2017.
- The question asks for the total number of medals won.
- India finished in the second position.
- The total number of medals secured by India was 24.

Revision Table: Jordan Junior and Cadet Open TT 2017

Event	Year	India's Finish	India's Total Medals
Jordan Junior and Cadet Open Table Tennis Tournament	2017	Second-best finisher	24

Additional Information: Table Tennis Tournaments

Table Tennis, also known as Ping Pong, is a racket sport that is popular worldwide. Junior and Cadet level tournaments are crucial for the development of young players, providing them with international exposure and competition experience.

- **International Table Tennis Federation (ITTF):** The governing body for table tennis worldwide.
- **Junior & Cadet Circuit:** ITTF organizes a circuit of events specifically for junior (<18 years) and cadet (<15 years) players. The Jordan Open was part of this circuit.

- **Importance of Junior Tournaments:** These events help identify and nurture future talent, offering ranking points that can help players qualify for bigger events like the World Junior Championships.
- **Indian Table Tennis Federation (TTFI):** The governing body for table tennis in India, responsible for selecting and sending teams to international events like the Jordan Open.

India's performance in such international junior events indicates the depth of talent in the country and the potential for future success at the senior level.

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66. **Answer: a**

Explanation:

Polly Umrigar Award and Indian Cricket Excellence

The Polly Umrigar Award is one of the most prestigious awards given by the Board of Control for Cricket in India (BCCI). It is presented to India's best international cricketer of the year. This award recognizes outstanding performance and contribution to Indian cricket on the international stage.

Identifying the Polly Umrigar Award Multiple Winner

The question asks which Indian cricketer received the Polly Umrigar Award and was the first Indian to win it on three separate occasions. Let's look at the options provided:

- Virat Kohli
- Sachin Tendulkar
- MS Dhoni
- Ravi Shastri

Historically, several great Indian cricketers have won the Polly Umrigar Award for their exceptional performances. However, achieving the milestone of winning it three times for the first time requires consistent excellence over multiple years.

Virat Kohli's Achievement with the Polly Umrigar Award

Among the listed cricketers, **Virat Kohli** holds the distinction mentioned in the question. He has won the Polly Umrigar Award multiple times, and he was indeed the first Indian cricketer to receive this prestigious award on three occasions. His consistent performance across formats during various periods earned him this recognition from the BCCI.

Winning the Polly Umrigar Award three times highlights Virat Kohli's dominance and impact as an international cricketer for India over a significant period.

Analysis of Other Options for the Polly Umrigar Award

Let's consider why the other options, while being legendary cricketers, do not fit the specific achievement asked in the question:

- **Sachin Tendulkar:** A legend of the game, Sachin Tendulkar also received the Polly Umrigar Award, but he was not the first to win it three times.
- **MS Dhoni:** A highly successful captain and player, MS Dhoni has also been a recipient of BCCI awards, but the specific record of being the first three-time Polly Umrigar Award winner belongs elsewhere.
- **Ravi Shastri:** A former player and coach, Ravi Shastri has been associated with Indian cricket for a long time, but his playing career period and award records differ from the achievement mentioned.

Based on the history of the Polly Umrigar Award and BCCI accolades, Virat Kohli is the cricketer who first achieved the feat of winning this award three times.

Revision Table: Key Polly Umrigar Award Winners

Cricketer	Notable Achievement related to Award
Virat Kohli	First Indian to win the Polly Umrigar Award three times.
Sachin Tendulkar	Recipient of the award.
MS Dhoni	Recipient of other significant BCCI awards; also received the Polly Umrigar Award.

Additional Information on BCCI Awards and Polly Umrigar Award

The Polly Umrigar Award is part of the annual BCCI Awards ceremony that recognizes and honors the top performers in Indian cricket. The award is named after Polly Umrigar, a distinguished former Indian captain and player. The criteria for the award typically focus on the player's performance in international matches during the preceding season.

Besides the Polly Umrigar Award for the best international cricketer, the BCCI also presents awards for best domestic players, best women cricketers, lifetime achievement awards, and others, celebrating excellence across all levels of Indian cricket.

Winning the Polly Umrigar Award is considered a significant honor, reflecting the player's contribution and impact on the team's success at the highest level.

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67. **Answer: c**

Explanation:

Understanding the Problem: Cost of Levelling Triangular Land

The question asks us to find the total cost required to level a piece of land shaped like a triangle. We are given the lengths of the three sides of the triangle and the cost to level per square metre.

To find the total cost, we first need to determine the area of the triangular land. Once the area is known, we can multiply it by the given rate of levelling per square metre to find the total cost.

Calculating the Area of the Triangular Land

The sides of the triangular land are given as 72 m, 30 m, and 78 m. Let's denote these sides as $a = 72 \text{ m}$, $b = 30 \text{ m}$, and $c = 78 \text{ m}$.

We can check if this triangle is a right-angled triangle using the Pythagorean theorem. The largest side is 78 m. Let's check if the sum of the squares of the other two sides equals the square of the largest side:

$$a^2 + b^2 = 72^2 + 30^2$$

$$a^2 + b^2 = 5184 + 900$$

$$a^2 + b^2 = 6084$$

Now, let's calculate the square of the longest side, c:

$$c^2 = 78^2$$

$$c^2 = 6084$$

Since $a^2 + b^2 = c^2$, the triangle is a right-angled triangle with the sides 30 m and 72 m being the base and height (or vice versa), and 78 m being the hypotenuse.

The area of a right-angled triangle is given by the formula:

$$\text{Area} = \frac{1}{2} \times \text{base} \times \text{height}$$

Using the two shorter sides as the base and height:

$$\text{Area} = \frac{1}{2} \times 30 \text{ m} \times 72 \text{ m}$$

$$\text{Area} = 15 \times 72 \text{ m}^2$$

$$\text{Area} = 1080 \text{ m}^2$$

Alternatively, we could have used Heron's formula for the area of a triangle with sides a, b, c and semi-perimeter $s = \frac{a+b+c}{2}$.

$$s = \frac{72 + 30 + 78}{2} = \frac{180}{2} = 90 \text{ m}$$

$$\text{Area} = \sqrt{s(s-a)(s-b)(s-c)}$$

$$\text{Area} = \sqrt{90(90-72)(90-30)(90-78)}$$

$$\text{Area} = \sqrt{90 \times 18 \times 60 \times 12}$$

$$\text{Area} = \sqrt{(9 \times 10) \times (2 \times 9) \times (6 \times 10) \times (2 \times 6)}$$

$$\text{Area} = \sqrt{9 \times 10 \times 2 \times 9 \times 6 \times 10 \times 2 \times 6}$$

Rearranging the terms to group squares:

$$\text{Area} = \sqrt{9^2 \times 10^2 \times 2^2 \times 6^2}$$

$$\text{Area} = \sqrt{(9 \times 10 \times 2 \times 6)^2}$$

$$\text{Area} = 9 \times 10 \times 2 \times 6$$

$$\text{Area} = 90 \times 12$$

$$\text{Area} = 1080 \text{ m}^2$$

Both methods confirm the area of the triangular land is 1080 square metres.

Calculating the Cost of Levelling

The rate of levelling is given as 20 paise per square metre. We need to convert this rate into Rupees per square metre.

Since 1 Rupee = 100 paise, 1 paise = $\frac{1}{100}$ Rupees.

So, 20 paise = $20 \times \frac{1}{100}$ Rupees = $\frac{20}{100}$ Rupees = 0.20 Rupees.

The rate is Rs. 0.20 per square metre.

The total cost of levelling is calculated by multiplying the area by the rate per square metre:

$$\text{Total Cost} = \text{Area} \times \text{Rate per square metre}$$

$$\text{Total Cost} = 1080 \text{ m}^2 \times \text{Rs. } 0.20/\text{m}^2$$

$$\text{Total Cost} = 1080 \times 0.20 \text{ Rupees}$$

$$\text{Total Cost} = 1080 \times \frac{20}{100} \text{ Rupees}$$

$$\text{Total Cost} = 1080 \times \frac{1}{5} \text{ Rupees}$$

$$\text{Total Cost} = \frac{1080}{5} \text{ Rupees}$$

$$\text{Total Cost} = 216 \text{ Rupees}$$

The cost of levelling the triangular piece of land is Rs. 216.

Summary of Calculations

Description	Value	Unit
Side 1	72	m
Side 2	30	m
Side 3	78	m
Semi-perimeter (s)	90	m
Area of Triangle	1080	m²
Levelling Rate (Paise)	20	Paise/m²
Levelling Rate (Rupees)	0.20	Rs./m²
Total Cost	216	Rs.

The calculated cost matches one of the provided options.

Revision Table: Area and Cost Calculation

Here's a quick review of the steps involved in calculating the area of a triangular piece of land and its levelling cost:

- Identify the sides of the triangular land.
- Determine the area of the triangle using an appropriate formula (e.g., Pythagorean theorem for right triangles, Heron's formula for scalene triangles).
- Ensure the area is in square metres.
- Convert the levelling rate to Rupees per square metre if given in paise.
- Multiply the area by the rate per square metre to find the total cost.

Additional Information: Triangular Land Area Formulas

The method to calculate the area of a triangular piece of land depends on the information available:

- **Base and Height:** If the base (b) and corresponding height (h) are known, Area $= \frac{1}{2} \times b \times h$.
- **Two Sides and Included Angle:** If two sides (a, b) and the angle (θ) between them are known, Area $= \frac{1}{2}ab\sin(\theta)$.
- **Three Sides (Heron's Formula):** If all three sides (a, b, c) are known, first calculate the semi-perimeter $s = \frac{a+b+c}{2}$, then Area $= \sqrt{s(s-a)(s-b)(s-c)}$. This formula is useful when the triangle is not right-angled or the height is not easily determined.
- **Right-Angled Triangle:** If the triangle is a right-angled triangle with legs (perpendicular sides) a and b, Area $= \frac{1}{2} \times \text{product of legs} = \frac{1}{2}ab$. This is a special case of the base and height formula where one leg is the base and the other is the height. We confirmed in this problem that the given sides form a right-angled triangle.

Converting currency units (like paise to Rupees) is crucial for accurate cost calculation problems.

68. **Answer: a**

Explanation:

This question asks us to analyze two statements and determine which of the two given conclusions logically follow from them. We will use logical deduction or Venn diagrams to visualize the relationships described in the statements and check if the conclusions hold true.

Understanding the Statements and Conclusions

Let's break down the given information:

Statements:

1. All blue are colours.
2. All colours are shades.

Conclusions:

1. All blue are shades.
2. Some shades are colours.

Analyzing the Logical Relationships

We can represent these statements using sets. Let B be the set of blue things, C be the set of colours, and S be the set of shades.

- Statement 1: "All blue are colours" means the set of blue things (B) is a subset of the set of colours (C). We can write this as $B \subseteq C$.
- Statement 2: "All colours are shades" means the set of colours (C) is a subset of the set of shades (S). We can write this as $C \subseteq S$.

Combining these two statements, we have $B \subseteq C$ and $C \subseteq S$. If set B is inside set C, and set C is inside set S, it logically follows that set B must also be inside set S.

Evaluating Conclusion 1: All blue are shades

Conclusion 1 states "All blue are shades". From our analysis of the statements ($B \subseteq C$ and $C \subseteq S$), we can deduce that $B \subseteq S$. This means that every element in set B is also an element in set S. In simple terms, if everything that is blue is a colour, and everything that is a colour is a shade, then it must be true that everything that is blue is also a shade.

Therefore, Conclusion 1 logically follows from the given statements.

Evaluating Conclusion 2: Some shades are colours

Conclusion 2 states "Some shades are colours". Statement 2 tells us that "All colours are shades" ($C \subseteq S$). If all colours are shades, it means that the set of colours (C) is entirely contained within the set of shades (S). If C is not an empty set (and it's reasonable to assume there are colours), then there are elements within the set S that are also in the set C. These elements are the colours themselves.

Saying "Some shades are colours" is equivalent to saying that the intersection of the set of shades (S) and the set of colours (C) is not empty, i.e., $S \cap C \neq \emptyset$. Since $C \subseteq S$ and we assume C is not empty, the intersection $S \cap C$ is exactly the set C itself, which is not empty. Thus, there are indeed some elements in S (shades) that are also in C (colours).

Therefore, Conclusion 2 logically follows from the given statements.

Summary of Conclusions

Based on our analysis:

- Conclusion 1: All blue are shades - Follows.
- Conclusion 2: Some shades are colours - Follows.

Both conclusions logically follow from the given statements.

Statement/Conclusion	Logical Representation	Follows?	Explanation
Statement 1: All blue are colours	$B \subseteq C$	N/A	Given statement
Statement 2: All colours are shades	$C \subseteq S$	N/A	Given statement
Conclusion 1: All blue are shades	$B \subseteq S$	Yes	Derived from $B \subseteq C$ and $C \subseteq S$
Conclusion 2: Some shades are colours	$S \cap C \neq \emptyset$ (since $C \subseteq S$)	Yes	If all colours are shades, then colours are a subset of shades, meaning some shades are colours (assuming colours exist).

Revision Table - Statements and Conclusions

Let's quickly review the relationship between universal affirmative statements ("All X are Y") and their implications.

Statement Type	Format	Venn Diagram	Key Implication
Universal Affirmative	All X are Y	Set X is inside Set Y ($X \subseteq Y$)	If Y is not empty, then Some Y are X ($Y \cap X \neq \emptyset$ or $X \subseteq Y \implies \text{Some Y are X if X is not empty}$)

In our case:

- Statement: All colours are shades ($C \subseteq S$).
- Implication: Some shades are colours (if colours exist). This supports Conclusion 2.

Also, the transitivity of subset relationships ($B \subseteq C$ and $C \subseteq S \implies B \subseteq S$) directly supports Conclusion 1.

Additional Information - Syllogisms and Deduction

This type of problem is a classic example of a categorical syllogism, which is a form of deductive reasoning. A syllogism consists of two premises (statements) and a conclusion. The validity of the conclusion depends solely on the logical form of the statements, not on the truthfulness of the content itself (assuming the statements are taken as true for the purpose of the argument).

The structure "All A are B" and "All B are C" leading to "All A are C" is a valid syllogism form (specifically, Barbara in the first figure). In our problem:

- All Blue (A) are Colours (B).
- All Colours (B) are Shades (C).
- Therefore, All Blue (A) are Shades (C). (Conclusion 1)

The conclusion "Some shades are colours" is derived from the relationship between the class 'Colours' and the class 'Shades' established in the statements. From "All colours are shades", it necessarily follows that there are items that are both shades and colours (namely, all the colours), provided the class 'colours' is not empty. In standard logic, existence is often assumed for the subject of universal statements when drawing existential conclusions, but modern interpretations sometimes handle this differently. However, in typical reasoning problems like this, "Some Y are X" is considered a valid deduction from "All X are Y" if X is assumed to exist.

Therefore, both conclusions logically follow from the statements provided.

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69. **Answer: d**

Explanation:

Understanding Pisces and Their Characteristics

The question asks us to identify which of the given options is NOT a characteristic feature of Pisces. Pisces is the group of animals commonly known as fish. They are aquatic vertebrates.

Key Characteristics of Pisces (Fish)

Let's examine the typical characteristics of animals belonging to the class Pisces:

- **Habitat:** Mostly aquatic, living in fresh or marine water.
- **Body Structure:** Streamlined body shape to facilitate movement in water.
- **Appendages:** Possess fins for swimming and balance.
- **Respiration:** Breathe using gills to extract dissolved oxygen from water.
- **Body Covering:** Body is usually covered with scales, which provide protection.
- **Skeleton:** Have an internal skeleton (endoskeleton) made of either bone (bony fish, Osteichthyes) or cartilage (cartilaginous fish, Chondrichthyes).
- **Body Temperature:** They are cold-blooded (poikilothermic or ectothermic), meaning their body temperature depends on the surrounding environment.
- **Circulatory System:** Have a two-chambered heart.
- **Body Cavity:** As vertebrates, they possess a true coelom (eucoelom).

Analyzing the Options

Let's evaluate each given option based on our knowledge of Pisces characteristics:

1. **Exoskeleton of scales:** Many fish have scales covering their body, which can be considered a type of exoskeleton providing protection. This is a common characteristic of Pisces.

2. **Breathing through gills:** Fish respire by using gills, which are specialized organs for gas exchange in water. This is a fundamental characteristic of Pisces.

3. **Endoskeleton of bone/cartilage:** Fish have an internal skeleton (endoskeleton). This skeleton is either made of bone (like in teleosts) or cartilage (like in sharks and rays). This is a defining characteristic of vertebrates, including Pisces.

4. **Presence of pseudocoelom:** A pseudocoelom is a body cavity that is not fully lined by mesoderm-derived tissue. It is found in certain invertebrates, such as

nematodes and rotifers. Vertebrates, including all members of Pisces, possess a true coelom (eucoelom), which is a body cavity completely lined by mesoderm. Therefore, the presence of a pseudocoelom is NOT a characteristic of Pisces.

Conclusion

Based on the analysis of the options and the characteristics of Pisces, the presence of a pseudocoelom is not a feature found in fish. Fish are vertebrates and have a true coelom.

Characteristics Analysis

Characteristic	Present in Pisces?	Explanation
Exoskeleton of scales	Yes (common)	Scales provide protection on the body surface.
Breathing through gills	Yes	Essential for respiration in water.
Endoskeleton of bone/cartilage	Yes	Internal support structure, made of bone or cartilage.
Presence of pseudocoelom	No	Pisces have a true coelom (eucoelom), not a pseudocoelom.

Revision Table: Key Pisces Characteristics

Characteristic	Description
Habitat	Aquatic (freshwater or marine)
Respiration	Gills
Body Covering	Usually scales
Skeleton	Endoskeleton of bone or cartilage
Body Temperature	Cold-blooded (Poikilothermic)
Body Cavity	True coelom (Eucoelom)

Additional Information: Types of Body Cavities

The presence and type of body cavity are important features used in animal classification. There are three main types:

- **Acoelomates:** Animals without a body cavity between the digestive tract and the outer body wall. Example: Flatworms (Platyhelminthes).
- **Pseudocoelomates:** Animals with a body cavity that is not fully lined by mesoderm. The body cavity lies between the mesoderm and endoderm. Example: Roundworms (Nematodes).
- **Eucoelomates (Coelomates):** Animals with a true coelom, which is a body cavity completely lined by mesoderm. This cavity is found between the body wall and the digestive tract. The mesoderm forms a lining called the peritoneum. All vertebrates, including Pisces, amphibians, reptiles, birds, and mammals, are eucoelomates.

Since Pisces are vertebrates, they possess a true coelom (eucoelom), not a pseudocoelom. This confirms why the presence of a pseudocoelom is not a characteristic of Pisces.

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70. Answer: c

Explanation:

Understanding the World Bank Loan for India's Waterway

The question asks about the specific loan amount sanctioned by The World Bank to support the development of India's first modern waterway. This refers to a significant infrastructure project aimed at improving connectivity and transportation.

Development projects, especially large-scale infrastructure initiatives like modern waterways, often require substantial investment. International financial institutions such as The World Bank play a crucial role in providing funding and technical assistance for such projects in developing countries like India.

The loan sanctioned by The World Bank was specifically designated to aid in the construction and development of the infrastructure required for this modern waterway, which facilitates inland water transport.

World Bank Sanctioned Amount for India's First Modern Waterway

Based on information regarding the funding for India's inland water transport development, The World Bank sanctioned a specific amount to support this initial modern waterway project.

The details of the loan are as follows:

- **Sanctioning Body:** The World Bank
- **Recipient Country:** India
- **Project Purpose:** Development of India's first modern waterway (specifically related to enhancing navigation and infrastructure on National Waterway 1, the Ganga River).
- **Loan Type:** Investment Project Financing

The sanctioned loan amount for this project was a specific figure aimed at covering various aspects of the waterway development, including fairway development, navigation aids, terminal construction, and institutional development.

Considering the options provided, the correct loan amount sanctioned by The World Bank for India's first modern waterway development project is $375 million.

Significance of the World Bank Loan for Waterway Development

The World Bank loan is significant for several reasons:

- It provides the necessary financial resources for a large-scale infrastructure project.
- It helps improve logistics and transportation efficiency in India.
- Inland waterways are often a more environmentally friendly and cost-effective mode of transport compared to road or rail for certain goods.
- The project contributes to regional development along the waterway corridor.
- It demonstrates international support for India's infrastructure goals.

This project, supported by the World Bank loan, focuses on making National Waterway 1 (NW1) a viable transport route, connecting important cities and industrial areas along the Ganga River.

Summary of World Bank Loan Details

Institution	Recipient	Project	Loan Amount
The World Bank	India	First Modern Waterway Development (NW1)	$375 million

Revision Table: Key Facts about India's Waterway Loan

Detail	Information
Funding Source	The World Bank
Purpose	Develop India's first modern waterway (NW1)
Sanctioned Amount	$375 million
Project Aim	Improve navigation, infrastructure on Ganga River

Additional Information on India's Waterway Infrastructure

India has a vast network of rivers, canals, backwaters, and creeks that can be used for inland water transport. The government is actively promoting the development of these waterways to reduce the burden on road and rail networks and provide a cheaper, more fuel-efficient, and eco-friendly mode of transport.

- National Waterway 1 (NW1) is the longest waterway in India, stretching from Prayagraj to Haldia.
- The project supported by the World Bank loan is part of the Jal Marg Vikas Project (JMVP) which aims to enhance navigation capacity of NW1.
- Developing waterways is crucial for connecting hinterlands to major ports and industrial hubs.
- Such infrastructure projects contribute to economic growth and regional connectivity.

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71. **Answer: c**

Explanation:

Understanding the Coin and Tumbler Experiment

The classic experiment where a coin placed on a card over a tumbler falls into the tumbler when the card is flicked demonstrates a fundamental principle in physics.

Let's break down why this happens.

The Principle: Law of Inertia

The phenomenon observed in the experiment is a direct consequence of the **Law of Inertia**. This law is also known as Newton's First Law of Motion.

Newton's First Law of Motion states:

- An object at rest stays at rest, and an object in motion stays in motion with the same speed and in the same direction unless acted upon by an unbalanced external force.

In simpler terms, objects resist changes to their state of motion. If they are not moving, they want to stay still. If they are moving, they want to keep moving in the same way.

Applying Inertia to the Coin and Tumbler

Consider the setup: a tumbler, a card covering its mouth, and a coin resting on the center of the card. Initially, both the coin and the card are at rest relative to the tumbler and the Earth.

- The coin is at rest. Due to its inertia, it tends to remain at rest.
- When the card is flicked horizontally with a finger, a force is applied to the card, causing it to move rapidly sideways.
- This flicking force acts primarily on the card, not directly on the coin (or at least, the force of friction between the coin and the card is minimal and acts only for a very short time).
- Because of its inertia, the coin resists this sudden horizontal motion. It tries to maintain its state of rest.
- As the card is quickly removed from under the coin, the coin loses the support it was resting on.
- Since no significant horizontal force acted on the coin to make it move sideways with the card, and it resisted the change due to inertia, it attempts to stay in its original horizontal position.
- However, there is a constant vertical force acting on the coin: gravity.

- With the support of the card gone, gravity pulls the coin downwards.
- As the coin was initially positioned directly over the opening of the tumbler, it falls vertically into the tumbler under the influence of gravity.

This experiment perfectly illustrates the concept of inertia – the property of an object to resist changes in its state of motion.

Why Other Options Are Not Primary Explanations

Let's briefly look at why the other options are not the primary explanation for this specific phenomenon:

- **Law of conservation of energy:** This law states that energy cannot be created or destroyed, only transformed. While energy transformations are involved (like the work done flicking the card), the fundamental reason the coin falls is inertia and gravity, not energy conservation principles explaining the motion itself in this specific context.
- **Newton's third law of motion:** This law deals with action-reaction pairs of forces (for every action, there is an equal and opposite reaction). While forces are present (e.g., friction between coin and card, force of flicking the card), this law doesn't directly explain why the coin remains in place horizontally and falls vertically.
- **Law of conservation of momentum:** This law states that the total momentum of an isolated system remains constant. Momentum is related to mass and velocity. While momentum is conserved in interactions, the falling of the coin is primarily explained by its initial state of rest and the lack of a horizontal force overcoming its inertia, combined with the vertical force of gravity. The conservation of momentum doesn't directly explain *why* the coin stays still horizontally.

Therefore, the **Law of Inertia** is the most accurate explanation for the coin falling into the tumbler.

Concept	Relevance to Coin/Tumbler
Law of Inertia (Newton's First Law)	Explains why the coin resists horizontal motion when the card is removed.
Newton's Second Law	Explains the coin's acceleration downwards due to gravity after support is removed ($F = ma$).
Gravity	The force that pulls the coin down into the tumbler once support is gone.
Newton's Third Law	Describes force pairs (e.g., coin on card, card on coin), but not the main principle for the fall.
Conservation of Energy/Momentum	Fundamental principles, but inertia is the most direct explanation for the coin staying put horizontally.

Revision Table: Key Physics Concepts

Concept	Brief Description	Associated Law(s)
Inertia	Resistance of an object to a change in its state of motion.	Newton's First Law of Motion
Force	A push or pull that can change an object's motion.	Newton's Laws of Motion
Gravity	The attractive force between objects with mass.	Newton's Law of Universal Gravitation, leads to weight ($W = mg$).
Momentum	The product of an object's mass and velocity ($p = mv$).	Law of Conservation of Momentum
Energy	The capacity to do work.	Law of Conservation of Energy

Additional Information on Inertia

Inertia is a property inherent in all objects with mass. The more mass an object has, the greater its inertia, and the more difficult it is to change its state of motion. For example, it's much harder to push a heavy box than a light one because the heavy box has more inertia. In the coin and tumbler experiment, the coin's mass gives it inertia, making it resist the horizontal movement of the card.

Inertia is often experienced daily:

- When a bus suddenly stops, passengers lean forward due to inertia, as their bodies tend to continue moving.
- When a bus suddenly starts, passengers are pushed backward as their bodies tend to remain at rest.

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72. **Answer: b**

Explanation:

Understanding the HCF and LCM Relationship

This question asks us to find one of two numbers given their Highest Common Factor (HCF), Least Common Multiple (LCM), and the value of the other number. We need to use the fundamental relationship between HCF, LCM, and the two numbers.

Key Property of HCF and LCM of Two Numbers

For any two positive integers, the product of the numbers is equal to the product of their HCF and LCM.

Mathematically, if the two numbers are a and b, then:

$a \times b = \text{HCF}(a, b) \times \text{LCM}(a, b)$

Applying the Property to Find the Other Number

We are given:

- HCF of the two numbers = 11

- LCM of the two numbers = 330
- One of the numbers = 55

Let the other number be x.

Using the property:

Product of the two numbers = HCF $\times$ LCM

$55 \times x = 11 \times 330$

Step-by-Step Calculation

We need to solve the equation $55 \times x = 11 \times 330$ for x.

To find x, we can divide the product of HCF and LCM by the given number:

$x = \frac{11 \times 330}{55}$

We can simplify the calculation:

$x = \frac{11 \times 330}{5 \times 11}$

Cancel out the common factor of 11:

$x = \frac{1 \times 330}{5}$

$x = \frac{330}{5}$

Now, perform the division:

$x = 66$

Finding the Other Number

The other number is 66.

Verification (Optional)

Let's check if the property holds with the two numbers 55 and 66:

Numbers are 55 and 66.

Product of numbers = $55 \times 66 = 3630$

HCF(55, 66): Factors of 55 are 1, 5, 11, 55. Factors of 66 are 1, 2, 3, 6, 11, 22, 33, 66. The highest common factor is 11. So, HCF = 11.

LCM(55, 66): We know HCF $\times$ LCM = 55×66. So, $11 \times \text{LCM} = 3630$. LCM = $\frac{3630}{11} = 330$.

Since HCF = 11 and LCM = 330, the property $55 \times 66 = 11 \times 330$ holds true ($3630 = 3630$). This confirms our answer is correct.

Given Information	Value
HCF of two numbers	11
LCM of two numbers	330
One number	55
Other number (let's call it x)	?

Revision Table: HCF and LCM Concepts

Concept	Definition	Key Property for Two Numbers a and b
HCF (Highest Common Factor)	The largest positive integer that divides both numbers without leaving a remainder. Also known as GCD (Greatest Common Divisor).	$a \times b = \text{HCF}(a, b) \times \text{LCM}(a, b)$
LCM (Least Common Multiple)	The smallest positive integer that is a multiple of both numbers.	

Additional Information on Finding HCF and LCM

Besides using the relationship with the product of the numbers, HCF and LCM can also be found using prime factorization:

- **Prime Factorization Method for HCF:** Express each number as a product of its prime factors. The HCF is the product of the common prime factors, each raised to the lowest power that appears in the factorizations.
- **Prime Factorization Method for LCM:** Express each number as a product of its prime factors. The LCM is the product of all prime factors (common and non-common), each raised to the highest power that appears in the factorizations.

For the numbers 55 and 66:

- Prime factorization of 55 = $5^1 \times 11^1$
- Prime factorization of 66 = $2^1 \times 3^1 \times 11^1$

- Common prime factor is 11. The lowest power of 11 is 11^1. So, HCF(55, 66) = 11.
- All prime factors are 2, 3, 5, 11. The highest power of 2 is 2^1, 3 is 3^1, 5 is 5^1, and 11 is 11^1. So, LCM(55, 66) = $2^1 \times 3^1 \times 5^1 \times 11^1 = 2 \times 3 \times 5 \times 11 = 6 \times 55 = 330$.

This method also confirms the HCF is 11 and the LCM is 330 for the numbers 55 and 66, reinforcing the relationship used to solve the problem.

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73. **Answer: c**

Explanation:

C → Mountain

A → Peak

B → Clouds

Every mountain has a peak.

Hence, the figure in option 3 is the correct answer.

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74. **Answer: c**

Explanation:

Understanding Syllogism Statements and Conclusions

This question asks us to analyze given statements and determine which of the provided conclusions logically follow from them. This type of problem falls under the category of logical reasoning, specifically syllogism.

Analyzing the Statements

We are given two statements:

1. All razors are blades.
2. All blades are metals.

We can represent these relationships conceptually. If something is a razor, it must be a blade. If something is a blade, it must be a metal.

Combining these statements, we can infer a relationship between razors and metals: Since all razors are blades, and all blades are metals, it logically follows that all razors must be metals.

Evaluating the Conclusions

Now let's examine each conclusion based on the given statements:

Conclusion 1: All metals are razors.

The statements tell us about the relationship in one direction: Razors are blades, and blades are metals. This means the set of razors is a subset of blades, and the set of blades is a subset of metals. However, the statements do not provide information about the reverse relationship. Just because all blades are metals doesn't mean everything that is a metal is also a blade. For example, gold is a metal but not a blade. Similarly, just because all razors are metals doesn't mean all metals are razors. There can be many metals that are not razors.

Therefore, conclusion 1, "All metals are razors," does not logically follow from the given statements.

Conclusion 2: Some metals are blades.

The second statement is "All blades are metals." This means that every single item that belongs to the category "blades" also belongs to the category "metals." If this is true, then the set of blades is entirely contained within the set of metals. If a set (blades) is contained within another set (metals), then at least some members of the larger set (metals) must also be members of the smaller set (blades). In fact, all the blades are metals.

So, if "All blades are metals" is true, then it is definitely true that "Some metals are blades." This conclusion logically follows from statement 2.

Determining Which Conclusions Follow

Based on our analysis:

- Conclusion 1: "All metals are razors" - Does NOT follow.
- Conclusion 2: "Some metals are blades" - DOES follow.

Therefore, only conclusion 2 logically follows from the given statements.

Summary of Analysis

Statement(s)	Analysis
All razors are blades.	Razors are a subset of Blades.
All blades are metals.	Blades are a subset of Metals.
Combined: Razors → Blades → Metals	Implies All razors are metals.

Conclusion	Follows?	Reasoning
1. All metals are razors.	No	Reverse relationship not guaranteed. Metals can exist that are not blades or razors.
2. Some metals are blades.	Yes	If all blades are metals, then the set of blades is inside the set of metals. Therefore, some members of the metal set are definitely blades.

The option that states "Only conclusion 2 follows" is the correct choice.

Revision Table: Key Concepts in Syllogism

Term	Explanation	Example
Statement	A premise providing a relationship between categories.	All A are B, Some B are C.
Conclusion	A deduction made based on the given statements.	Some A are C (if logically derived).
"All A are B"	Means the set of A is entirely within the set of B.	Set A $\subset$ Set B
"Some A are B"	Means there is at least one element common to sets A and B.	Set A $\cap$ Set B $\neq \emptyset$
Logical Inference	Drawing necessary conclusions from true premises.	If "All dogs are mammals" and "All mammals are animals", then "All dogs are animals" follows.

Additional Information: Syllogism Rules and Types

Syllogisms are a fundamental part of deductive reasoning. They consist of statements (premises) and a conclusion. The goal is to determine if the conclusion is necessarily true given that the premises are true.

- **Categorical Syllogism:** This type, used in the question, deals with statements about categories or classes (like razors, blades, metals) using terms like "all," "no," "some," and "some not."
- **Quantifiers:** "All" is a universal affirmative quantifier. "Some" is a particular affirmative quantifier.
- **Subset Relationship:** The statement "All A are B" establishes a subset relationship where A is a subset of B. From this, it always follows that "Some B are A" (assuming A is not empty, which is standard in these problems) and "Some A are B" (which is simply restating part of the original premise).

- **Drawing Conclusions:** Conclusions must follow necessarily. If there is any possible scenario where the statements are true but the conclusion is false, the conclusion does not logically follow.

Understanding the relationships implied by "all" and "some" is crucial for solving these logical reasoning problems.

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75. **Answer: a**

Explanation:

Calculating Percentage Increase in CO_2 Emissions

The question asks us to find the percentage increase in CO_2 emissions from the year 2016 to the year 2017, based on the data provided in the table.

Year	Domestic CO_2 emission (million metric tons)
2015	100
2016	110
2017	150

From the table, we can see the CO_2 emission values for the relevant years:

- CO_2 emission in 2016 = 110 million metric tons
- CO_2 emission in 2017 = 150 million metric tons

Step-by-Step Calculation

To calculate the percentage increase, we first need to find the absolute increase in CO_2 emissions from 2016 to 2017.

Step 1: Calculate the absolute increase in CO_2 emission.

Increase = CO2 emission in 2017 - CO2 emission in 2016

Increase = $150 - 110 = 40$ million metric tons

So, the CO2 emission increased by 40 million metric tons from 2016 to 2017.

Step 2: Calculate the percentage increase.

The formula for percentage increase is:

$$\text{Percentage Increase} = \left(\frac{\text{Increase}}{\text{Original Value}}\right) \times 100\%$$

In this case, the original value is the emission in the starting year, which is 2016.

Original Value = CO2 emission in 2016 = 110 million metric tons

Substitute the values into the formula:

$$\text{Percentage Increase} = \left(\frac{40}{110}\right) \times 100\%$$

Now, let's simplify the fraction and perform the calculation:

$$\frac{40}{110} = \frac{4}{11}$$

$$\text{Percentage Increase} = \left(\frac{4}{11}\right) \times 100\%$$

$$\text{Percentage Increase} \approx 0.363636... \times 100\%$$

$$\text{Percentage Increase} \approx 36.3636...\%$$

Rounding to two decimal places, the percentage increase is approximately 36.36%.

This calculation shows that the increase in the percentage of CO2 emission from 2016 to 2017 is 36.36%.

Revision Table: Key Concepts

Concept	Description
Percentage Change	A way to express how much a quantity changes relative to its original value, as a fraction of 100.
Percentage Increase	Calculated when the new value is greater than the original value. Formula: $\left(\frac{\text{New Value - Original Value}}{\text{Original Value}}\right) \times 100\%$.
CO2 Emissions	The release of carbon dioxide into the atmosphere, often measured in metric tons.

Additional Information on Percentage Increase

Understanding percentage change is a crucial skill in interpreting data like CO2 emissions or economic indicators. A percentage increase tells us not just the raw amount of increase, but how significant that increase is compared to where it started. For example, an increase of 10 million tons might seem small if the original emission was 1000 million tons (1% increase), but very large if the original emission was only 20 million tons (50% increase).

When calculating percentage increase:

- Identify the original value (the starting point).
- Identify the new value (the ending point).
- Calculate the difference (Increase = New Value - Original Value).
- Divide the increase by the original value.
- Multiply the result by 100 to convert it to a percentage.

It's important to always use the original value as the denominator in the fraction. Using the new value or the difference in the denominator would result in an incorrect percentage change calculation.

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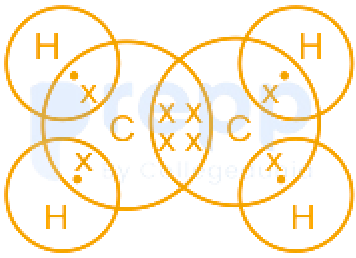
76. **Answer: c**

Explanation:

The Correct answer is**Option Fig. 3**

- The electronic dot and cross structure of **ethene** is:



- As we know that, the molecular formula of Ethene is C_2H_4. Hence, the above diagram shows the correct dot and cross structure of Ethene.
- **Ethylene/Ethene** is an unsaturated hydrocarbon.
- It is a colorless flammable gas with a faint sweet and musky gas.
- It is naturally produced and released by most fresh fruits and vegetables as a natural ripening agent.
- It is also known as the**natural ripening hormone**.

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77. **Answer: d**

Explanation:

Understanding the Least Common Multiple (LCM)

The Least Common Multiple (LCM) of a set of numbers is the smallest positive integer that is a multiple of all the numbers in the set. Finding the LCM is useful in many mathematical problems, such as adding or subtracting fractions with different denominators.

Finding LCM using Prime Factorization

One common and effective way to find the LCM of a set of numbers is by using the prime factorization method. This involves the following steps:

1. Find the prime factorization of each number in the set.
2. For each prime factor, identify the highest power that appears in any of the factorizations.
3. Multiply these highest powers of all the prime factors together. The result is the LCM.

Prime Factorization of the Given Numbers

We need to find the LCM of 16, 24, 36, 52, and 54. Let's find the prime factorization for each number:

- Prime factorization of 16:

  $$16 = 2 \times 2 \times 2 \times 2 = 2^4$$

- Prime factorization of 24:

  $$24 = 2 \times 2 \times 2 \times 3 = 2^3 \times 3^1$$

- Prime factorization of 36:

  $$36 = 2 \times 2 \times 3 \times 3 = 2^2 \times 3^2$$

- Prime factorization of 52:

  $$52 = 2 \times 2 \times 13 = 2^2 \times 13^1$$

- Prime factorization of 54:

  $$54 = 2 \times 3 \times 3 \times 3 = 2^1 \times 3^3$$

Identifying Highest Powers of Prime Factors

Now, let's look at all the prime factors we found (2, 3, and 13) and identify the highest power for each across the factorizations:

- For the prime factor 2: The powers are 2^4 (from 16), 2^3 (from 24), 2^2 (from 36), 2^2 (from 52), and 2^1 (from 54). The highest power of 2 is $\mathbf{2^4}$.

- For the prime factor 3: The powers are 3^1 (from 24), 3^2 (from 36), and 3^3 (from 54). There is no factor of 3 in 16 and 52, which can be considered 3^0. The highest power of 3 is $\mathbf{3^3}$.
- For the prime factor 13: The power is 13^1 (from 52). There is no factor of 13 in the other numbers, which can be considered 13^0. The highest power of 13 is $\mathbf{13^1}$.

Calculating the LCM

To find the LCM, we multiply the highest powers of all prime factors:

$\text{LCM}(16, 24, 36, 52, 54) = (\text{Highest power of } 2) \times (\text{Highest power of } 3) \times (\text{Highest power of } 13)$

$\text{LCM} = 2^4 \times 3^3 \times 13^1$

Let's calculate the values:

- $2^4 = 16$
- $3^3 = 3 \times 3 \times 3 = 27$
- $13^1 = 13$

So, $\text{LCM} = 16 \times 27 \times 13$

First, multiply 16 by 27:

$16 \times 27 = 432$

Now, multiply 432 by 13:

$432 \times 13 = 432 \times (10 + 3) = (432 \times 10) + (432 \times 3)$

$432 \times 10 = 4320$

$432 \times 3 = 1296$

$4320 + 1296 = 5616$

The LCM of 16, 24, 36, 52, and 54 is 5616.

Number	Prime Factorization
16	2^4
24	$2^3 \times 3^1$
36	$2^2 \times 3^2$
52	$2^2 \times 13^1$
54	$2^1 \times 3^3$

Highest powers: 2^4, 3^3, 13^1

LCM = $2^4 \times 3^3 \times 13^1 = 16 \times 27 \times 13 = 5616.$

Revision Table: LCM Calculation Steps

Step	Action	Details
1	Prime Factorize	Break down each number (16, 24, 36, 52, 54) into its prime factors.
2	List Factors & Powers	Write down the prime factorization for each number, showing the powers of the prime factors.
3	Identify Highest Powers	For every unique prime factor found (2, 3, 13), determine the highest power that appears in any of the factorizations.
4	Calculate LCM	Multiply these highest powers together to get the LCM.

Additional Information: LCM vs. HCF

It's helpful to distinguish LCM from Highest Common Factor (HCF), also known as Greatest Common Divisor (GCD).

- **LCM**: The smallest positive integer divisible by all numbers in the set. Found by taking the **highest** power of each prime factor.

- **HCF/GCD**: The largest positive integer that divides into all numbers in the set without leaving a remainder. Found by taking the **lowest** power of each **common** prime factor.

For the numbers 16, 24, 36, 52, 54:

- Prime factors involved: 2, 3, 13
- Only common prime factor is 2.
- Lowest power of 2 is 2^1 (from 54).
- No other prime factors (3, 13) are common to ALL numbers.
- So, HCF = $2^1 = 2$.

This shows how LCM (5616) and HCF (2) are calculated differently using prime factorizations.

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78. **Answer: b**

Explanation:

Understanding Materials Used for Lenses

A lens is a transparent optical device that refracts light rays to either converge or diverge them. Lenses are fundamental components in many optical instruments like cameras, telescopes, microscopes, and eyeglasses. For a material to be used in making lenses, it must possess certain key properties:

- **Transparency:** Light must be able to pass through the material clearly with minimal scattering or absorption.
- **Ability to Refract Light:** The material must have a refractive index different from the surrounding medium (usually air) so that it can bend light rays as they pass through.
- **Homogeneity:** The material's composition and density should be uniform throughout to ensure predictable and consistent refraction.
- **Ability to be Shaped:** The material must be able to be molded, ground, or polished into the required curved surfaces.

Analyzing the Given Options for Lens Materials

Let's examine each option provided to determine which material is unsuitable for making lenses:

- **Plastic:** Many types of plastic (e.g., acrylic, polycarbonate) are transparent, can be molded into complex shapes, and have suitable refractive indices. Plastic lenses are commonly used, especially for eyeglasses due to their lightness and impact resistance. So, plastic can be used for making lenses.
- **Soil:** Soil is typically an opaque or translucent mixture of mineral particles, organic matter, water, and air. It is not transparent, it is not homogeneous, and it cannot be shaped into a clear, smooth optical surface that would effectively refract light in a controlled manner. Therefore, soil cannot be used for making lenses.
- **Glass:** Various types of glass (e.g., crown glass, flint glass) are the traditional and most common materials for making high-quality lenses. Glass is transparent, homogeneous, has well-defined refractive properties, and can be precisely ground and polished into lens shapes. So, glass can be used for making lenses.
- **Water:** While not used to make solid, permanent lenses, water can act as a lens. A drop of water on a surface can act as a simple magnifying lens. Water itself is transparent and refracts light. Principles of water acting as a lens are sometimes used in specific applications or as demonstrations (e.g., liquid lenses). So, water can, under certain conditions, be considered for making lenses or lens-like structures.

Conclusion on Unsuitable Material for Lenses

Based on the properties required for lens materials and the analysis of the options, soil is the material that absolutely cannot be used for making lenses because it lacks the essential properties of transparency, homogeneity, and the ability to form a smooth, predictable optical surface required for refracting light effectively.

The material that cannot be used for making lenses among the given options is Soil.

Revision Table: Lens Material Suitability

Material	Transparency	Homogeneity	Can be Shaped	Suitable for Lenses?
Plastic	Yes	Yes (typically)	Yes	Yes
Soil	No (Opaque/Translucent)	No	No (cannot form smooth optical surface)	No
Glass	Yes	Yes (typically)	Yes	Yes
Water	Yes	Yes (typically pure water)	Yes (can form shapes via surface tension)	Yes (as liquid lenses)

Additional Information: Properties of Lens Materials

The effectiveness of a lens material depends on its optical properties, particularly its refractive index and dispersion. Different types of glass and plastic are developed with specific refractive indices and dispersion characteristics to correct aberrations (like chromatic aberration) in optical systems.

- **Refractive Index (n):** This value indicates how much light is bent when passing from one medium to another. A higher refractive index means light is bent more. Lenses require materials with a significantly different refractive index than the surrounding air ($n \approx 1$).
- **Dispersion:** This refers to how the refractive index varies with the wavelength of light. Materials with low dispersion are preferred to minimize chromatic aberration, where different colors of light are focused at slightly different points.

Materials like soil, being non-transparent and non-uniform, cannot be characterized by a specific refractive index or dispersion in a way that would allow them to function as a predictable lens.

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79. **Answer: c**

Explanation:

In the given figures all are mathematical operators except D which has an alphabet.

Hence, figure D is the odd one.

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80. **Answer: d**

Explanation:

Maximum number of mobile phones is produced by company LMN

Number of mobile phones produced = $1240000 \times 30/100 = 372000$

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81. **Answer: c**

Explanation:

Understanding the Transformative Chief Minister Award

The question asks about a specific award, the 'Transformative Chief Minister Award', presented by the US-India Business Council (USIBC) in May 2017. This award is given to recognize Indian Chief Ministers who have made significant contributions through policy changes and reforms to improve the business environment and attract investment in their respective states.

Identifying the Award Recipient in May 2017

In May 2017, the US-India Business Council (USIBC) presented the 'Transformative Chief Minister Award' during its India Investment Summit in New York. This recognition was bestowed upon a Chief Minister for their leadership in implementing reforms and fostering economic growth.

Let's consider the options provided:

- K. Chandrashekhar Rao: Was the Chief Minister of Telangana during this period.
- Nitish Kumar: Was the Chief Minister of Bihar during this period.
- N. Chandrababu Naidu: Was the Chief Minister of Andhra Pradesh during this period.
- Yogi Adityanath: Became the Chief Minister of Uttar Pradesh in March 2017.

Researching the recipients of the USIBC 'Transformative Chief Minister Award' for May 2017 reveals that the award was presented to N. Chandrababu Naidu, the then Chief Minister of Andhra Pradesh. He was recognized for his efforts in promoting business reforms, ease of doing business, and attracting investments to the state.

Conclusion on the Transformative Chief Minister Awardee

Based on the information available regarding the US-India Business Council (USIBC) awards presented in May 2017, the 'Transformative Chief Minister Award' was given to N. Chandrababu Naidu for his impactful work in Andhra Pradesh.

Chief Ministers and States in May 2017 (Contextual)

Chief Minister	State (in May 2017)
K. Chandrashekhar Rao	Telangana
Nitish Kumar	Bihar
N. Chandrababu Naidu	Andhra Pradesh
Yogi Adityanath	Uttar Pradesh

Therefore, among the given options, N. Chandrababu Naidu is the Chief Minister who received the 'Transformative Chief Minister Award' from the US-India Business Council (USIBC) in May 2017.

Revision Table: Key Details

Award Details for Revision

Award Name	Presenting Body	Year	Recipient Chief Minister	State
Transformative Chief Minister Award	US-India Business Council (USIBC)	May 2017	N. Chandrababu Naidu	Andhra Pradesh

Additional Information: USIBC and State Reforms

The US-India Business Council (USIBC) plays a vital role in strengthening economic ties between the United States and India. It often recognizes leaders in India who contribute significantly to improving the business and investment climate, making it easier for both domestic and international companies to operate. Awards like the 'Transformative Chief Minister Award' highlight the importance of state-level reforms in driving India's overall economic growth and attracting foreign direct investment (FDI).

Chief Ministers who focus on simplifying regulations, improving infrastructure, ensuring policy stability, and promoting specific sectors are often lauded for their efforts in creating a favorable environment for business and industry.

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82. **Answer: c**

Explanation:

Calculating Principal Amount with Simple Interest

The question asks us to find the principal amount that would earn a simple interest of Rs. 120 at an annual interest rate of 6% over a period of 10 years. We can use the formula for simple interest to solve this problem.

Simple Interest Formula

The formula for calculating Simple Interest (SI) is:

$SI = \frac{P \times R \times T}{100}$

Where:

- P = Principal amount (the initial sum of money)
- R = Rate of interest per annum
- T = Time period in years
- SI = Simple Interest earned

Given Values from the Question

From the problem statement, we are given the following information:

- Simple Interest (SI) = Rs. 120
- Rate of Interest (R) = 6% per annum
- Time Period (T) = 10 years
- Principal (P) = ? (This is what we need to find)

Finding the Principal (P)

To find the principal amount (P), we need to rearrange the simple interest formula. We can do this by multiplying both sides of the formula by 100 and then dividing by $R \times T$:

$SI \times 100 = P \times R \times T$

$P = \frac{SI \times 100}{R \times T}$

Step-by-Step Calculation of Principal

Now, let's substitute the given values into the rearranged formula:

$P = \frac{120 \times 100}{6 \times 10}$

First, calculate the product of rate and time in the denominator:

$6 \times 10 = 60$

Next, calculate the product of simple interest and 100 in the numerator:

$120 \times 100 = 12000$

Now, substitute these values back into the formula for P:

$P = \frac{12000}{60}$

Perform the division:

$P = 200$

So, the principal amount is Rs. 200.

Verification

Let's verify if a principal of Rs. 200 at 6% p.a. for 10 years yields Rs. 120 simple interest:

$SI = \frac{200 \times 6 \times 10}{100}$

$SI = \frac{12000}{100}$

$SI = 120$

This matches the given simple interest, confirming our calculated principal is correct.

Parameter	Value
Simple Interest (SI)	Rs. 120
Rate (R)	6%
Time (T)	10 years
Principal (P)	Rs. 200

Therefore, the principal that will yield Rs. 120 as simple interest at 6% p.a. in 10 years is Rs. 200.

Revision Table: Simple Interest Concepts

Term	Definition	Formula (if applicable)
Principal (P)	The initial amount of money borrowed or invested.	$P = \frac{SI \times 100}{R \times T}$
Rate (R)	The percentage at which interest is calculated per year.	$R = \frac{SI \times 100}{P \times T}$
Time (T)	The duration for which the money is borrowed or invested.	$T = \frac{SI \times 100}{P \times R}$
Simple Interest (SI)	Interest calculated only on the principal amount.	$SI = \frac{P \times R \times T}{100}$
Amount (A)	The total sum of Principal and Simple Interest.	$A = P + SI$ or $A = P\left(1 + \frac{R \times T}{100}\right)$

Additional Information on Simple Interest

Simple interest is the most basic form of interest calculation. It is always calculated on the original principal amount, regardless of how much interest has accrued in previous periods. This differs from compound interest, where interest is calculated on the principal amount plus any accumulated interest from previous periods. Simple interest is commonly used for short-term loans or for basic savings accounts.

Key points about simple interest calculation:

- The rate of interest is usually given annually (per annum, p.a.). If given otherwise (e.g., half-yearly, quarterly), it must be converted to an annual rate or the time period adjusted accordingly.
- The time period must be in years. If given in months or days, it must be converted into years (e.g., 6 months = $\frac{6}{12}$ years, 73 days = $\frac{73}{365}$ years).
- Simple interest remains constant for each year if the principal and rate are unchanged.

83. **Answer: a**

Explanation:

In figure 2 only the first triangle of figure 1 changes its orientation.

Similarly,

In figure 4 the first triangle of figure 3 will change its orientation.

.Hence, figure C is the correct answer

84. **Answer: a**

Explanation:

Finding the Curved Surface Area of a Right Circular Cone

This problem asks us to find the curved surface area of a right circular cone given its height and slant height. We are provided with the height (**h**) as 24 cm and the slant height (**l**) as 25 cm. We are also told to use the value of π as 22/7.

To find the curved surface area (CSA) of a cone, we use the formula:

$$\text{CSA} = \pi \times \text{radius} \times \text{slant height} = \pi r l$$

However, we are not directly given the radius (**r**). In a right circular cone, the height, radius, and slant height form a right-angled triangle. The slant height is the hypotenuse of this triangle. Therefore, we can use the Pythagorean theorem to find the radius:

$$\text{height}^2 + \text{radius}^2 = \text{slant height}^2$$

$$h^2 + r^2 = l^2$$

Calculating the Radius of the Cone

We have the height **h** = 24 cm and the slant height **l** = 25 cm. Let's substitute these values into the Pythagorean theorem:

$$(24 \text{ cm})^2 + r^2 = (25 \text{ cm})^2$$

Now, let's calculate the squares:

$$576 \text{ cm}^2 + r^2 = 625 \text{ cm}^2$$

To find r^2, subtract 576 from 625:

$$r^2 = 625 \text{ cm}^2 - 576 \text{ cm}^2$$

$$r^2 = 49 \text{ cm}^2$$

Take the square root of both sides to find the radius **r**:

$$r = \sqrt{49 \text{ cm}^2}$$

$$r = 7 \text{ cm}$$

So, the radius of the base of the cone is 7 cm.

Calculating the Curved Surface Area of the Cone

Now that we have the radius (**r** = 7 cm), the slant height (**l** = 25 cm), and the value of π (22/7), we can calculate the curved surface area using the formula $\text{CSA} = \pi r l$:

$$\text{CSA} = \frac{22}{7} \times 7 \text{ cm} \times 25 \text{ cm}$$

We can cancel out the 7 in the numerator and the denominator:

$$\text{CSA} = 22 \times 25 \text{ cm}^2$$

Now, multiply 22 by 25:

$$22 \times 25 = 22 \times (20 + 5) = 22 \times 20 + 22 \times 5 = 440 + 110 = 550$$

So, the curved surface area is:

$$\text{CSA} = 550\ \text{cm}^2$$

The curved surface area of the right circular cone is 550 cm².

Summary of the Calculation

- Given: Height (h) = 24 cm, Slant height (l) = 25 cm, π = 22/7.
- Find radius (r) using $h^2 + r^2 = l^2$: $24^2 + r^2 = 25^2 \implies 576 + r^2 = 625 \implies r^2 = 49 \implies r = 7$ cm.
- Calculate Curved Surface Area (CSA) using $\text{CSA} = \pi r l$: $\text{CSA} = \frac{22}{7} \times 7 \times 25 = 22 \times 25 = 550$ cm².

Comparing our result with the given options, we find that 550 cm² matches one of the choices.

Quantity	Value	Units
Height (h)	24	cm
Slant Height (l)	25	cm
Radius (r)	7	cm
Curved Surface Area (CSA)	550	cm²

Revision Table: Right Circular Cone Formulas

Concept	Formula	Notes
Pythagorean Relation	$h^2 + r^2 = l^2$	Relates height, radius, and slant height.
Curved Surface Area (CSA)	$\text{CSA} = \pi r l$	Area of the slanted surface.
Base Area	$\text{Base Area} = \pi r^2$	Area of the circular base.
Total Surface Area (TSA)	$\text{TSA} = \text{CSA} + \text{Base Area} = \pi r l + \pi r^2 = \pi r(l + r)$	Total area including the base.
Volume (V)	$V = \frac{1}{3}\pi r^2 h$	Space occupied by the cone.

Additional Information: Properties of a Right Circular Cone

A right circular cone is a three-dimensional geometric shape that tapers smoothly from a flat circular base to a point called the apex or vertex. The apex is located directly above the center of the base. Key features include:

- **Base:** A perfect circle.
- **Height (h):** The perpendicular distance from the apex to the center of the base.
- **Radius (r):** The distance from the center of the base to any point on its circumference.
- **Slant Height (l):** The distance from the apex to any point on the circumference of the base, measured along the surface of the cone.
- The angle between the height and the slant height is a right angle (90°).

Understanding these properties and the relationships between height, radius, and slant height is crucial for solving problems involving the surface area and volume of a right circular cone.

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85. **Answer: b**

Explanation:

Understanding the Best Film on Social Issues Award at the 64th National Film Awards

The question asks about the recipient of the National Award for Best Film on Social Issues at the 64th National Film Awards, which were presented in 2017 for films released in 2016. This award category recognizes films that effectively address significant social issues and contribute to public awareness or inspire positive change.

Analyzing the Options for the 64th National Film Awards

Let's examine the given options in the context of the 64th National Film Awards:

- **Dangal:** This movie was a critically acclaimed and commercially successful film based on the lives of the Phogat family and their daughters' wrestling careers. While it addressed themes of gender equality and overcoming societal norms, it did not win the National Award specifically for Best Film on Social Issues at the 64th awards.
- **Pink:** This film dealt with the critical issue of consent and women's rights in India. Its narrative strongly highlighted social problems faced by women. It was widely praised for its powerful message and performances.
- **Sultan:** This movie was a sports drama about a wrestler. While popular, it was not primarily focused on deep social issues in the way the National Award category defines.
- **Airlift:** Based on a true story of the evacuation of Indians from Kuwait during the Gulf War, this film was a patriotic drama. It focused on a historical event rather than contemporary social issues targeted by the specific award category.

Identifying the Winner: 64th National Film Awards

Based on the results of the 64th National Film Awards, the movie that was recognized for its significant contribution to highlighting social issues was 'Pink'. The film's strong stand on consent and its portrayal of the challenges faced by women made it a fitting choice for the Best Film on Social Issues award.

Conclusion on the National Film Awards Winner

Therefore, among the given options, 'Pink' is the film that received the National Award for Best Film on Social Issues at the 64th National Film Awards held in 2017.

Movie Title	Consideration for Best Film on Social Issues (64th National Film Awards)
Dangal	Addressed social themes but not the winner of this specific award.
Pink	Won the National Award for Best Film on Social Issues.
Sultan	Not awarded in this category.
Airlift	Not awarded in this category.

Revision Table: 64th National Film Awards - Key Winner

Award Category	Winning Film (64th National Film Awards, 2017)
Best Film on Social Issues	Pink

Additional Information: National Film Awards and Social Issues

The National Film Awards are among the most prestigious film awards in India, recognizing excellence in cinematic achievements. The category of 'Best Film on Social Issues' is particularly important as it encourages filmmakers to use the powerful medium of cinema to shed light on pressing societal problems, spark conversations, and potentially contribute to societal change. Films awarded in this category often tackle sensitive or overlooked issues, bringing them to the forefront of public discourse. The 64th National Film Awards celebrated a diverse range of Indian cinema, and the recognition of 'Pink' underscored the importance of addressing issues like consent and women's safety through mainstream cinema.

86. **Answer: b**

Explanation:

Analyzing the Group: Animal Classification

The question asks us to identify the animal that does not belong to the group consisting of Fox, Goat, Horse, and Zebra.

Let's examine each animal and consider common ways to group animals:

- **Fox:** Foxes are mammals. They are typically classified as carnivores (meat-eaters) or omnivores (eating both plants and meat).
- **Goat:** Goats are mammals. They are known herbivores, meaning they primarily eat plants.
- **Horse:** Horses are mammals. They are herbivores, primarily grazing on grasses.
- **Zebra:** Zebras are mammals. Like horses, they are herbivores, feeding mainly on grasses.

Considering their diets provides a clear distinction:

- Goats are herbivores.
- Horses are herbivores.
- Zebras are herbivores.
- Foxes are not primarily herbivores; they eat meat and sometimes plants.

Therefore, based on their dietary habits, the Fox stands out from the other animals in the group.

Another possible classification could be based on whether they are typically domesticated or wild. Goats and Horses are commonly domesticated, while Zebras and Foxes are typically wild. This grouping doesn't clearly isolate one animal.

However, the difference in diet (herbivore vs. carnivore/omnivore) is a fundamental biological classification often used to group animals. This distinction clearly separates the Fox from the Goat, Horse, and Zebra.

Hence, the animal that does NOT belong to the group based on this characteristic is the Fox.

Revision Table: Animal Grouping

Animal	Typical Diet Classification	Belongs to Herbivore Group?
Fox	Carnivore/Omnivore	No
Goat	Herbivore	Yes
Horse	Herbivore	Yes
Zebra	Herbivore	Yes

Additional Information: Types of Animal Diets

Understanding different types of animal diets is a key concept in biology and ecology. The main classifications are:

- **Herbivores:** Animals that primarily eat plants. Examples include cows, deer, rabbits, horses, and zebras. Their digestive systems are often adapted to break down tough plant material.
- **Carnivores:** Animals that primarily eat meat. Examples include lions, tigers, wolves, and some types of fish. They typically have sharp teeth and claws for hunting and tearing flesh. Foxes are often included here as they primarily hunt smaller animals, although they are sometimes classified as omnivores.
- **Omnivores:** Animals that eat both plants and meat. Examples include humans, bears, pigs, and many birds. They have a more varied diet and digestive system adapted to process both types of food. Foxes are sometimes considered omnivores because they may supplement their diet with fruits, berries, or insects.
- **Detritivores:** Animals that eat dead organic matter (detritus). Examples include earthworms, millipedes, and some insects.

These dietary classifications help define an animal's role in an ecosystem's food chain.

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87. **Answer: b**

Explanation:

Understanding the Series Pattern: YC3, XF6, WI9

The question asks us to find the next term in the given series: YC3, XF6, WI9, ______. To solve this type of series question, we need to examine the pattern in each component of the terms: the first letter, the second letter, and the number.

Analyzing the Components of the Series

Let's break down each term into its components and observe the sequence:

Term	First Letter	Second Letter	Number
YC3	Y	C	3
XF6	X	F	6
WI9	W	I	9

Pattern in the First Letter

The sequence of the first letters is Y, X, W. Let's look at their positions in the English alphabet:

- Y is the 25th letter.
- X is the 24th letter.
- W is the 23rd letter.

We can see a clear pattern here: the first letter is decreasing by one position each time.

The difference between consecutive first letters is:

- X is 1 position before Y.
- W is 1 position before X.

Following this pattern, the next first letter should be the letter that is 1 position before W. The letter before W is V.

Pattern in the Second Letter

The sequence of the second letters is C, F, I. Let's look at their positions in the English alphabet:

- C is the 3rd letter.
- F is the 6th letter.
- I is the 9th letter.

Let's find the difference in positions between consecutive second letters:

- F is $6 - 3 = 3$ positions after C.
- I is $9 - 6 = 3$ positions after F.

The pattern here is that the second letter is increasing by three positions each time.

Following this pattern, the next second letter should be the letter that is 3 positions after I. The 9th letter is I. The 12th letter is L.

Pattern in the Number

The sequence of the numbers is 3, 6, 9. Let's look at the difference between consecutive numbers:

- $6 - 3 = 3$
- $9 - 6 = 3$

The pattern here is that the number is increasing by 3 each time. This is a simple arithmetic progression with a common difference of 3.

Following this pattern, the next number should be $9 + 3 = 12$.

Combining the Patterns

Based on our analysis of each component, the next term in the series should have:

- First letter: V
- Second letter: L
- Number: 12

Combining these, the next term in the series is VL12.

Conclusion

By analyzing the individual patterns within the series YC3, XF6, WI9, we determined the next term. The first letter decreases by one step in the alphabet, the second letter increases by three steps, and the number increases by three. Applying these rules, the next term is VL12.

Revision Table: Series Analysis

Component	Sequence	Pattern	Next Element
First Letter	Y, X, W	Decreasing by 1 letter position	V
Second Letter	C, F, I	Increasing by 3 letter positions	L
Number	3, 6, 9	Increasing by 3	12

Additional Information: Types of Letter and Number Series

Series completion questions often involve identifying patterns in letters, numbers, or a combination of both. Common types of patterns include:

- **Arithmetic Progression:** Numbers increase or decrease by a constant difference (like the number series 3, 6, 9).
- **Geometric Progression:** Numbers increase or decrease by a constant ratio (multiplication or division).
- **Alphabetical Series:** Letters follow a pattern based on their position in the alphabet, either sequential, skipping letters, or based on positional values.
- **Mixed Series:** A combination of patterns for different components, as seen in this problem where letters follow positional patterns and numbers follow an arithmetic progression.
- **Fibonacci Series:** Each number is the sum of the two preceding ones (e.g., 1, 1, 2, 3, 5, 8...).
- **Skipping Patterns:** Elements in the series skip a fixed number of items in a sequence (e.g., skipping every alternate letter).

Solving these requires careful observation and breaking down complex terms into simpler components.

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88. Answer: a

Explanation:

1) From the figure, $\angle ACB = 180° - (72° + 69°) = 180° - 141° = 39°$

According to the question,

$\Rightarrow x° + 39° = 180°$

$\Rightarrow x° = 141°$

$\therefore$ From (1) we get the value of x

2) $\angle ACB = 180 - x°$

Here, we do not get value of x, using only (2)

$\therefore$ Only 1 is sufficient

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89. **Answer: a**

Explanation:

After folding, the sheet will look like the below figure:

Hence, the figure given in option figure C is the correct answer.

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90. **Answer: c**

Explanation:

Understanding the 90th Oscar Best Picture Award

The Academy Awards, also known as the Oscars, are prestigious awards presented annually by the Academy of Motion Picture Arts and Sciences (AMPAS) to recognize excellence in cinematic achievements. The "Best Picture" award is the most significant award given, recognizing the best film of the year.

The question asks about the movie that won the **Best Picture award** at the **90th Oscar awards**. This event took place on March 4, 2018, honouring films released in 2017.

To answer this question, we need to recall or look up the results of the **90th Academy Awards**, specifically the winner of the **Best Picture** category.

The nominees for the **Best Picture** award at the **90th Oscar awards** included several acclaimed films. The winner is selected through a preferential ballot system by members of the Academy.

Let's examine the options provided:

- Phantom Thread
- Three Billboards Outside Ebbing, Missouri (often referred to as Three Billboards)
- The Shape of Water
- Coco

Among these nominated films (and other nominees in the category), the film that ultimately received the highly coveted **Best Picture** trophy at the **90th Oscar awards** was The Shape of Water.

Here is a summary of the nominees for the **Best Picture award** at the **90th Oscars**:

Movie Title	Result
Call Me By Your Name	Nominee
Darkest Hour	Nominee
Dunkirk	Nominee
Get Out	Nominee
Lady Bird	Nominee
Phantom Thread	Nominee
The Post	Nominee
The Shape of Water	**Winner**
Three Billboards Outside Ebbing, Missouri	Nominee

Therefore, The Shape of Water won the **Best Picture award** at the **90th Oscar awards**.

Revision Table: Key Details on 90th Oscar Best Picture

Award Ceremony	Award Category	Winning Movie
90th Academy Awards (Oscars)	Best Picture	The Shape of Water

Additional Information on the Shape of Water and the 90th Oscars

The Shape of Water is a fantasy drama film directed by Guillermo del Toro. It tells the story of a mute cleaning lady at a high-security government laboratory in 1962 who falls in love with a captured amphibious creature.

Besides winning **Best Picture** at the **90th Oscar awards**, The Shape of Water also won other major awards that night, including Best Director for Guillermo del Toro, Best Original Score, and Best Production Design. It was the film with the most nominations (13) and the most wins (4) at the **90th Oscars**.

Understanding the winners and nominees of major awards like the **Oscar Best Picture** award is important for general knowledge and film history studies.

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91. **Answer: d**

Explanation:

Understanding Forces Inside the Atom's Nucleus

The question asks about the force between a neutron and a proton inside an atom. To answer this, we need to consider the fundamental forces at play within the tiny confines of the atomic nucleus where protons and neutrons reside.

Analyzing the Options

Let's look at the forces listed in the options:

1. Tidal force: This force arises from the difference in gravitational pull across a body due to another celestial body. It is relevant on astronomical scales, not

within an atom.

2. Gravitational force: Gravity exists between any two objects with mass. Protons and neutrons have mass, so there is a gravitational attraction between them. However, gravity is the weakest of the fundamental forces, and its strength at the atomic scale is negligible compared to the forces that hold the nucleus together or push it apart.
3. Electrostatic force: This force exists between charged particles. Protons are positively charged $(+e)$, and neutrons are electrically neutral (0). There is an electrostatic force between two protons (repulsive), but no electrostatic force directly between a neutron and a proton because the neutron has no net charge. While electrostatic repulsion between protons is present in the nucleus, it's not the force binding neutrons and protons.
4. Nuclear force: This refers to the strong nuclear force, which is the fundamental force that binds quarks and gluons together to form protons and neutrons, and the residual strong force (often just called the nuclear force) which holds protons and neutrons (collectively called nucleons) together in the nucleus of an atom. This force is strongly attractive at typical nucleon distances within the nucleus.

The Role of the Nuclear Force

Inside the nucleus, protons (positive charge) repel each other strongly via the electrostatic force. Without a powerful attractive force to counteract this repulsion, the nucleus would fly apart. The gravitational force is far too weak to do this.

The force that holds protons and neutrons together in the nucleus is the strong nuclear force. This force has some key properties:

- It is a very **strong attractive force** at short distances (about 10^{-15} meters), which is the typical separation between nucleons in a nucleus.
- It is **short-range**, meaning its strength drops off very rapidly beyond nuclear distances.
- It is **charge-independent**, meaning the attractive force between a proton and a proton $(p\text{-}p)$, a neutron and a neutron $(n\text{-}n)$, and a proton and a neutron $(p\text{-}n)$ is approximately the same at the same separation, ignoring the electrostatic repulsion between protons.

Therefore, the force that a neutron exerts on a proton inside an atom, which helps keep the nucleus stable, is the nuclear force.

Force	Applies Between	Nature at Nuclear Scale
Tidal Force	Large masses (celestial bodies)	Irrelevant in atoms
Gravitational Force	Objects with mass	Weak attraction; negligible in nucleus stability
Electrostatic Force	Charged particles	Repulsive between protons; none directly p-n
Nuclear Force	Protons and Neutrons (Nucleons)	Strong, short-range attraction

Conclusion

Based on the nature and strength of the fundamental forces within the atomic nucleus, the force responsible for binding neutrons and protons is the nuclear force.

Revision Table: Forces within the Atom

Force Type	Involved Particles	Relative Strength in Nucleus	Range
Strong Nuclear Force	Quarks (binds nucleons); Nucleons (binds nucleus)	Strongest	Short ($\sim 10^{-15}$ m)
Electromagnetic Force	Charged particles (Protons, Electrons)	Strong (but weaker than nuclear force in nucleus)	Infinite
Weak Nuclear Force	Elementary particles (involved in decay)	Very Weak	Very Short ($\sim 10^{-18}$ m)
Gravitational Force	Particles with mass (Protons, Neutrons, Electrons)	Weakest	Infinite

Additional Information: The Strong Force and Nuclear Stability

The nuclear force is a manifestation of the more fundamental strong force, which binds quarks together to form protons and neutrons. The residual strong force acts between these composite particles (nucleons).

The balance between the attractive nuclear force and the repulsive electrostatic force between protons determines the stability of an atomic nucleus. For larger nuclei, more neutrons are needed to provide enough nuclear attraction (which acts between all nucleons) to overcome the increasing total electrostatic repulsion between protons (which only acts between protons).

Understanding these forces is crucial in nuclear physics and explains phenomena like radioactive decay and nuclear reactions.

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92. **Answer: b**

Explanation:

Understanding Pipe Filling Rates

This problem involves understanding the concept of work rate, specifically how quickly pipes can fill a tank. The rate of work is the amount of work done per unit of time. In this case, the 'work' is filling the tank, and the 'rate' is the fraction of the tank filled per hour.

Relating the Speeds of Pipes X, Y, and Z

We are given relationships between the speeds (rates) of the three pipes:

- Pipe Y is twice as fast as pipe X. This means if X fills a certain amount in an hour, Y fills double that amount in the same time.
- Pipe Z is thrice as fast as pipe Y. This means if Y fills a certain amount in an hour, Z fills triple that amount in the same time.

Let's express the rates in terms of a common variable. Let the rate of pipe X be R_X (fraction of tank filled per hour).

- Rate of pipe Y, R_Y, is twice the rate of X: $R_Y = 2 \times R_X$
- Rate of pipe Z, R_Z, is thrice the rate of Y: $R_Z = 3 \times R_Y = 3 \times (2 \times R_X) = 6 \times R_X$

So, the rates of the three pipes are in the ratio $R_X : R_Y : R_Z = R_X : 2R_X : 6R_X = 1 : 2 : 6$.

Calculating the Combined Filling Rate

When the pipes X, Y, and Z work together, their rates add up. The combined rate is the sum of their individual rates.

Combined Rate $R_{XYZ} = R_X + R_Y + R_Z$

Substituting the rates in terms of R_X:

$R_{XYZ} = R_X + 2R_X + 6R_X = 9R_X$

The combined rate of filling the water tank is 9 times the rate of pipe X.

Time Taken and Rate Relationship

The time taken to fill the tank is inversely proportional to the filling rate. If a pipe or set of pipes fills the tank at a rate R (fraction per hour), the time taken to fill the whole tank (which is 1 unit of work) is $Time = \frac{1}{Rate}$.

We are given that pipes X, Y, and Z together fill the water tank in 5 hours.

This means their combined rate is $\frac{1}{5}$ of the tank per hour.

So, $R_{XYZ} = \frac{1}{5}$.

Finding the Time for Pipe X Alone

We know that $R_{XYZ} = 9R_X$ and $R_{XYZ} = \frac{1}{5}$.

Therefore, we can set up the equation:

$$9R_X = \frac{1}{5}$$

Now, we need to find the time taken by pipe X alone. The time taken by pipe X alone, let's call it T_X, is related to its rate R_X by the formula $T_X = \frac{1}{R_X}$.

From the equation $9R_X = \frac{1}{5}$, we can solve for R_X:

$$R_X = \frac{1}{9 \times 5} = \frac{1}{45}$$

The rate of pipe X is $\frac{1}{45}$ of the tank per hour.

Now, we find the time taken by pipe X alone:

$$T_X = \frac{1}{R_X} = \frac{1}{1/45} = 45 \text{ hours.}$$

So, pipe X alone will take 45 hours to fill the water tank.

Step-by-Step Calculation

Here is a breakdown of the steps:

1. Assign a rate to pipe X: Let Rate of X = R.

2. Determine the rates of Y and Z based on the given information:
    - Rate of Y = 2 * Rate of X = $2R$
    - Rate of Z = 3 * Rate of Y = 3 * (2R) = $6R$
3. Calculate the combined rate of X, Y, and Z:
    - Combined Rate = Rate of X + Rate of Y + Rate of Z = $R + 2R + 6R = 9R$
4. Determine the combined rate from the total time taken (5 hours):
    - Combined Rate = $\frac{1}{\text{Time taken together}} = \frac{1}{5}$ tank/hour
5. Equate the two expressions for the combined rate:
    - $9R = \frac{1}{5}$
6. Solve for R, the rate of pipe X:
    - $R = \frac{1}{9 \times 5} = \frac{1}{45}$ tank/hour
7. Calculate the time taken by pipe X alone using its rate:
    - Time for X = $\frac{1}{\text{Rate of X}} = \frac{1}{R} = \frac{1}{1/45} = 45$ hours

The time taken by pipe X alone to fill the water tank is 45 hours.

Revision Table: Pipe Filling Time and Rate

Pipe	Rate (in terms of R_X)	Time to fill tank alone (if $R_X = 1/T_X$)
X	R_X	T_X
Y	$2R_X$	$T_X/2$
Z	$6R_X$	$T_X/6$
X, Y, Z together	$9R_X$	$T_X/9$

We are given that the time taken by X, Y, Z together is 5 hours. From the table, this time is $T_X/9$. So, $T_X/9 = 5$. Solving for T_X: $T_X = 5 \times 9 = 45$ hours.

Additional Information: Concepts in Pipes and Cisterns

Problems involving pipes filling or emptying tanks are similar to time and work problems. Here are some key concepts:

- **Work Rate:** The amount of work (filling or emptying a fraction of the tank) done per unit of time. It's usually expressed as "fraction of tank filled per hour" or "fraction of tank emptied per hour".
- **Filling Pipes:** These pipes do positive work (add water to the tank). Their rates are positive.
- **Emptying Pipes (Leaks):** These pipes do negative work (remove water from the tank). Their rates are negative.
- **Combined Rate:** When multiple pipes work simultaneously, their individual rates are added algebraically (filling rates are positive, emptying rates are negative).
- **Relationship between Rate and Time:** If a pipe or a combination of pipes can complete a job (fill the tank) in T hours, their combined rate is $\frac{1}{T}$ of the tank per hour. Conversely, if the rate is R per hour, the time taken is $\frac{1}{R}$ hours.
- **Total Work:** The total work is usually considered as 1 unit (representing the full tank).

Understanding these basics is crucial for solving problems related to pipes and cisterns, especially those involving different rates and combined work.

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93. **Answer: c**

Explanation:

Understanding the Location of Garo Hills

The question asks about the geographical location of the Garo Hills in India. The options provided are different states in the northeastern part of the country.

The Garo Hills are a part of a larger hill range system in Northeast India, known for their natural beauty and distinct tribal culture, primarily inhabited by the Garo people.

Where are the Garo Hills Situated?

The Garo Hills are predominantly located in the state of Meghalaya. Meghalaya is known as the "Abode of the Clouds" and is famous for its hills, valleys, lakes, and caves. The state's geography is defined by three major hill ranges running from west to east: the Garo Hills, the Khasi Hills, and the Jaintia Hills.

The Garo Hills form the western part of this prominent plateau region.

Analyzing the Options

Let's look at the given options:

1. Assam: Assam is a neighboring state to Meghalaya, located primarily in the Brahmaputra Valley. While there might be foothills or bordering areas, the main Garo Hills range is not located in Assam.
2. Nagaland: Nagaland is located further east in Northeast India, known for the Naga Hills. The Garo Hills are geographically separate from Nagaland.
3. Meghalaya: As discussed, the Garo Hills constitute the western part of Meghalaya's plateau. This is where the major portion of the Garo Hills lies.
4. Mizoram: Mizoram is located in the southern part of Northeast India, known for the Mizo Hills. It is geographically distant from the Garo Hills.

Based on geographical facts, the Garo Hills are located in Meghalaya.

Conclusion on Garo Hills Location

The Garo Hills are an important geographical feature of Northeast India. Their location within Meghalaya is a fundamental aspect of the state's geography and cultural landscape.

Garo Hills Location Summary

Geographical Feature	Primary Location (State)
Garo Hills	Meghalaya
Khasi Hills	Meghalaya
Jaintia Hills	Meghalaya
Naga Hills	Nagaland
Mizo Hills	Mizoram

Revision Table: Garo Hills Location Facts

Key Facts about Garo Hills

Detail	Description
Location	Primarily in the state of Meghalaya, India.
Part of	The Meghalaya Plateau, western section.
Inhabited by	Mainly the Garo people.
Neighboring Hills	Khasi Hills to the east.

Additional Information: Meghalaya's Hill Ranges and Geography

Meghalaya's terrain is dominated by a plateau that is segmented into the western Garo Hills, the central Khasi Hills, and the eastern Jaintia Hills. These hills are part of the older Indian Peninsular Shield.

- The Garo Hills are less elevated than the Khasi Hills.
- Tura is a major town located in the Garo Hills.
- The region receives very high rainfall, contributing to its lush green landscape and numerous waterfalls.

- Together, the Garo, Khasi, and Jaintia Hills define the state's physiography and influence its climate and biodiversity.

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94. **Answer: d**

Explanation:

The question asks about a significant approval under the Pradhan Mantri Awas Yojana (PM Awas Yojana), specifically concerning the urban poor. This scheme is a major initiative by the Indian government aimed at providing housing for all.

Understanding the PM Awas Yojana

The Pradhan Mantri Awas Yojana (PMAY) is a flagship housing scheme launched by the Government of India. It has two main components:

- **Pradhan Mantri Awas Yojana (Urban) - PMAY-U:** Aims to address the housing requirement of the urban poor, including slum dwellers.
- **Pradhan Mantri Awas Yojana (Gramin) - PMAY-G:** Focuses on providing housing assistance to the rural poor.

The scheme provides central assistance to Urban Local Bodies (ULBs) and other implementing agencies through States/UTs for vertical slum rehabilitation, affordable housing in partnership, beneficiary-led individual house construction/enhancement, and credit-linked subsidy schemes.

Approval of 20 Lakh Houses for Urban Poor

Approvals under the PM Awas Yojana (Urban) are typically made by the Ministry responsible for Housing and Urban Affairs. At the time relevant to the question, Shri Venkaiah Naidu held the position of the Union Minister for Housing and Urban Affairs, as well as Information and Broadcasting. He was instrumental in the implementation and approvals under the PM Awas Yojana (Urban).

It was during his tenure as the Minister of Housing and Urban Affairs that the approval for 20 lakh houses for the urban poor under the PM Awas Yojana (Urban) was made. This was a significant step towards achieving the mission's goal of 'Housing for All'.

Analyzing the Options

Let's look at the provided options:

1. Smriti Irani: While a prominent minister, her portfolio has typically included Women & Child Development, Textiles, and Education, not Urban Housing.
2. Arun Jaitley: He was the Finance Minister, responsible for the budget and economic affairs, not directly the Minister for Urban Housing.
3. Sushma Swaraj: She was primarily the Minister of External Affairs.
4. Shri Venkaiah Naidu: As discussed, he was the Union Minister for Housing and Urban Affairs when such approvals under PM Awas Yojana (Urban) were made.

Based on the responsibilities and portfolios of the individuals listed at the relevant time, Shri Venkaiah Naidu was the minister responsible for approving housing projects under the PM Awas Yojana (Urban).

Therefore, Shri Venkaiah Naidu approved the 20 lakh houses for urban poor under the PM Awas Yojana.

Revision Table: Key Points on PM Awas Yojana

Scheme	Objective	Components
Pradhan Mantri Awas Yojana (PMAY)	Housing for All by 2022 (original target)	PMAY-Urban, PMAY-Gramin
PMAY-Urban	Housing for urban poor	Slum rehabilitation, Affordable Housing in Partnership, Beneficiary-led construction, Credit-linked subsidy

Additional Information: PM Awas Yojana (Urban) Approvals

The approval of housing units under PM Awas Yojana (Urban) involves various stages and authorities. Significant milestones, such as the approval of a large number of houses like 20 lakh, are typically announced by the Union Minister in charge of the Ministry of Housing and Urban Affairs. These approvals are based on proposals submitted by States and Union Territories, which consolidate plans from Urban Local Bodies. The scheme aims to cover beneficiaries from Economically Weaker Sections (EWS) and Lower Income Groups (LIG) in urban areas, providing financial assistance to make housing affordable.

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95. **Answer: d**

Explanation:

Understanding the Question: First Woman Lawyer Appointed to Supreme Court

The question asks about a specific milestone in the history of the Indian judiciary: the first woman lawyer who was directly appointed as a judge to the Supreme Court of India. It's important to note the key phrase "directly appointed as a judge", which distinguishes this path from judges who are elevated to the Supreme Court from High Courts.

Process of Appointing Supreme Court Judges

Judges of the Supreme Court of India can be appointed through two main processes:

- **Elevation from High Courts:** Most Supreme Court judges are appointed by being elevated from their positions as judges (usually Chief Justices) in the High Courts of India.

- **Direct Appointment from the Bar:** Lawyers practicing in the Supreme Court or High Courts (referred to as the Bar) can also be directly appointed as judges of the Supreme Court, provided they meet certain eligibility criteria, such as being a distinguished jurist or having practiced for a specified number of years. This is a less common route compared to elevation from High Courts.

Identifying the First Woman Lawyer Directly Appointed

Historically, there have been several esteemed women judges in the Supreme Court, but most were elevated from High Courts. The question specifically looks for the first woman who took the less common path of direct appointment from the Bar.

Let's consider the options provided:

- Vrinda Grover: Known lawyer and human rights activist.
- Menaka Guruswamy: Senior Advocate practicing in the Supreme Court.
- Meenakshi Arora: Senior Advocate practicing in the Supreme Court.
- Indu Malhotra: Senior Advocate who practiced in the Supreme Court.

Among these options, Smt. Indu Malhotra holds the distinction of being the first woman lawyer to be directly appointed as a judge of the Supreme Court of India.

Indu Malhotra's Historic Appointment

Indu Malhotra was a senior advocate and became the first woman to be directly appointed from the Bar to the Supreme Court. She was appointed as a Supreme Court judge in April 2018. Prior to her appointment, she had a distinguished career as a lawyer, specializing in arbitration law.

This appointment marked a significant moment, as it broke the barrier for women lawyers being directly elevated to the highest court of the country, alongside judges elevated from High Courts.

Comparing with Other Notable Women Judges

It's worth noting that the first woman judge in the Supreme Court was Justice M. Fathima Beevi, appointed in 1989. However, she was elevated from the Kerala High Court, not directly appointed from the Bar. Several other women judges have since been appointed to the Supreme Court, predominantly through elevation from High Courts, making Indu Malhotra's direct appointment from the Bar a unique first.

Conclusion: The First Woman Lawyer to be Directly Appointed

Based on the history of appointments to the Supreme Court of India, Smt. Indu Malhotra is the first woman lawyer who was directly appointed as a judge from the Bar.

Revision Table: Supreme Court Appointments

Criteria	Description	Example (First Woman)
First Woman Judge (overall)	First woman appointed to the Supreme Court, regardless of appointment method.	Justice M. Fathima Beevi (Elevated from High Court)
First Woman Judge Directly from Bar	First woman lawyer appointed directly to the Supreme Court from legal practice.	Justice Indu Malhotra (Directly from Bar)

Additional Information: Supreme Court of India

The Supreme Court of India is the apex judicial body in the country. It consists of the Chief Justice of India and a maximum of 34 judges (including the Chief Justice). Judges are appointed by the President of India after consultation with such of the judges of the Supreme Court and of the High Courts in the States as the President may deem necessary for the purpose.

Eligibility criteria for appointment as a Supreme Court judge include:

- Citizen of India.
- Has been a Judge of a High Court for at least five years; OR
- Has been an Advocate of a High Court or of two or more such Courts in succession for at least ten years; OR
- Is, in the opinion of the President, a distinguished jurist.

The appointment of Justice Indu Malhotra fulfilled the criterion of having been an Advocate for at least ten years (specifically, she was a Senior Advocate at the Supreme Court).

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96. **Answer: b**

Explanation:

Understanding Electric Charge, Current, and Time

In physics, electric current is defined as the rate of flow of electric charge. This relationship is fundamental to understanding how electricity behaves in a circuit. The amount of electric charge (denoted by Q) that flows through a circuit is directly proportional to both the electric current (denoted by I) and the time (denoted by t) for which the current flows.

The formula connecting these three quantities is:

$$Q = I \times t$$

Where:

- Q is the electric charge in Coulombs (C)
- I is the electric current in Amperes (A)
- t is the time in seconds (s)

Calculating Electric Charge: Step-by-Step

The question provides us with the values for the electric current and the time for which the current flows through the circuit:

- Current (I) = 0.6 A
- Time (t) = 6 minutes

Before we can use the formula $Q = I \times t$, we need to ensure that all units are in the standard SI system. The current is already in Amperes, which is the standard unit for current. However, the time is given in minutes, and the standard unit for time in this formula is seconds.

Converting Time to Seconds

There are 60 seconds in 1 minute. So, to convert the time from minutes to seconds, we multiply the number of minutes by 60:

$$t(\text{in seconds}) = \text{Time (in minutes)} \times 60$$

$$t = 6 \text{ minutes} \times 60 \frac{\text{seconds}}{\text{minute}}$$

$$t = 360 \text{ seconds}$$

Applying the Formula to Find Electric Charge

Now that we have the current in Amperes and the time in seconds, we can use the formula $Q = I \times t$ to calculate the electric charge:

$$Q = I \times t$$

Substitute the given values:

$$Q = 0.6 \text{ A} \times 360 \text{ s}$$

$$Q = 216 \text{ C}$$

Therefore, the amount of electric charge flowing through the circuit is 216 Coulombs.

Summary of Calculation

Quantity	Symbol	Given Value	Standard Unit	Value in Standard Unit
Electric Current	I	0.6 A	Ampere (A)	0.6 A
Time	t	6 minutes	second (s)	$6 \times 60 = 360$ s
Electric Charge	Q	?	Coulomb (C)	$Q = I \times t = 0.6 \times 360 = 216$ C

The calculated amount of electric charge is 216 C.

Revision Table: Key Electrical Concepts

Concept	Definition	Formula (Basic)	SI Unit
Electric Charge (Q)	A fundamental property of matter that causes it to experience a force when placed in an electromagnetic field.	$Q = I \times t$	Coulomb (C)
Electric Current (I)	The rate of flow of electric charge past a point or region.	$I = \frac{Q}{t}$	Ampere (A)
Time (t)	Duration over which the charge flows.	$t = \frac{Q}{I}$	second (s)

Additional Information on Electric Charge and Current

- **Definition of Ampere:** One Ampere is defined as the flow of one Coulomb of charge per second. This directly comes from the relationship $I = Q/t$.
- **Definition of Coulomb:** One Coulomb is the amount of charge transported in one second by a current of one Ampere. This comes from $Q = I \times t$. It is a very

large amount of charge; the charge of a single electron is approximately 1.602×10^{-19} C.

- **Types of Current:** Current can be direct current (DC), where charge flows in one direction, or alternating current (AC), where the direction of charge flow reverses periodically. The formula $Q = It$ is most straightforwardly applied to a constant DC current over time t. For varying currents, integration might be needed.
- **Conservation of Charge:** Electric charge is a conserved quantity. This means that the total charge in an isolated system remains constant.

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97. **Answer: b**

Explanation:

Finding the Missing Number in the Series

The question asks us to identify the missing number in the given series: 27, 29, 33, ?, 57. To find the missing number, we need to analyze the pattern or rule that governs the sequence of numbers.

Analyzing the Number Series Pattern

Let's look at the differences between consecutive terms in the series:

- Difference between the second and first term:

$$29 - 27 = 2$$

- Difference between the third and second term:

$$33 - 29 = 4$$

We observe that the first difference is 2 and the second difference is 4. It appears that the differences are increasing, possibly following a specific mathematical sequence.

Identifying the Pattern Rule

The differences found are 2 and 4. Let's consider potential patterns for these differences:

- **Arithmetic Progression of Differences:** If the differences formed an arithmetic progression, the next difference would be

$$4 + (4 - 2) = 4 + 2 = 6$$

.

- **Geometric Progression of Differences:** If the differences formed a geometric progression, the next difference would be

$$4 \times (4/2) = 4 \times 2 = 8$$

.

- **Powers of a Number:** The differences 2 and 4 could be powers of 2 (

$$2 = 2^1$$

,

$$4 = 2^2$$

). If this pattern continues, the next difference would be

$$2^3 = 8$$

.

Let's test the "powers of 2" pattern for the differences.

- Term 2 = Term 1 +

$$2^1 = 27 + 2 = 29$$

(Correct)

- Term 3 = Term 2 +

$$2^2 = 29 + 4 = 33$$

(Correct)

- If the pattern continues, the missing number (Term 4) should be Term 3 +

$$2^3$$

.
- The next term (Term 5) should be Term 4 +

$$2^4$$

.

Calculating the Missing Number

Using the identified pattern where the difference is the next power of 2:

- Missing Number (Term 4) = Term 3 +

$$2^3$$

- Missing Number =

$$33 + 8$$

- Missing Number =

$$41$$

Verifying the Pattern with the Last Term

Now, let's check if adding the next power of 2 (

$$2^4$$

) to the calculated missing number gives the last term (57):

- Term 5 = Missing Number +

$$2^4$$

- Term 5 =

$$41 + 16$$

- Term 5 =

$$57$$

(Correct)

The pattern holds true for the entire series. The differences between consecutive terms are

$$2^1, 2^2, 2^3, 2^4$$

.

Terms	Value	Difference from previous term	Pattern of Difference
Term 1	27	-	-
Term 2	29	$29 - 27 = 2$	2^1
Term 3	33	$33 - 29 = 4$	2^2
Term 4 (Missing)	41	$41 - 33 = 8$	2^3
Term 5	57	$57 - 41 = 16$	2^4

Thus, the missing number in the series 27, 29, 33, ?, 57 is 41.

Revision Table: Number Series Analysis

Concept	Description	Application in this problem
Number Series	A sequence of numbers following a specific pattern.	Given series 27, 29, 33, ?, 57.
Pattern Recognition	Identifying the rule connecting consecutive terms.	Examining differences between terms (2, 4).
Difference Series	A new series formed by the differences between consecutive terms of the original series.	The difference series starts with 2, 4. Pattern identified as powers of 2.
Solving for Missing Term	Using the identified pattern to calculate the unknown value.	Adding 2^3 (which is 8) to 33 to find the missing term.

Additional Information: Types of Number Series

Number series problems in reasoning and quantitative aptitude often involve different types of patterns. Some common types include:

- **Arithmetic Series:** Each term is obtained by adding a constant value to the previous term.
- **Geometric Series:** Each term is obtained by multiplying the previous term by a constant value.
- **Difference Series:** The differences between consecutive terms follow a pattern (e.g., arithmetic, geometric, or another sequence like prime numbers, squares, cubes, etc.). This is the type seen in this problem.
- **Ratio Series:** The ratio between consecutive terms follows a pattern.
- **Mixed Series:** Involves a combination of two or more patterns (e.g., alternating addition and subtraction, or combining arithmetic and geometric operations).
- **Fibonacci Series:** Each term is the sum of the two preceding terms (usually starting with 0 and 1 or 1 and 1).

Solving number series problems requires careful observation, calculation of differences or ratios, and testing various common patterns.

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98. **Answer: b**

Explanation:

Understanding Profit and Loss Percentage Calculation

This problem involves calculating the cost price (CP) of an article based on a given selling price (SP) and profit percentage, and then determining the gain or loss percentage when the article is sold at a different selling price.

Step 1: Calculate the Cost Price (CP)

We are given that selling the article for Rs. 595 results in a 19% gain. We can use the formula relating Selling Price (SP), Cost Price (CP), and Profit Percentage (P%):

$SP = CP \times \left(1 + \frac{P\%}{100}\right)$

Plugging in the given values:

$595 = CP \times (1 + \frac{19}{100})$

$595 = CP \times (1 + 0.19)$

$595 = CP \times 1.19$

Now, we can solve for CP:

$CP = \frac{595}{1.19}$

To make the division easier, we can multiply the numerator and denominator by 100:

$CP = \frac{595 \times 100}{1.19 \times 100} = \frac{59500}{119}$

Performing the division:

$CP = 500$

So, the cost price of the article is Rs. 500.

Step 2: Determine Gain or Loss at the New Selling Price

The new selling price (SP2) is given as Rs. 475. We compare this new selling price with the calculated cost price (CP = Rs. 500).

Since $SP2 = 475$ and $CP = 500$, we see that $SP2 < CP$.

When the selling price is less than the cost price, there is a loss.

Step 3: Calculate the Loss Amount

The loss amount is the difference between the cost price and the selling price:

Loss = CP - SP2

Loss = $500 - 475$

Loss = Rs. 25

Step 4: Calculate the Loss Percentage

The loss percentage is calculated with respect to the cost price using the formula:

$\text{Loss percentage (L\%)} = \left(\frac{\text{Loss}}{\text{CP}}\right) \times 100$

Plugging in the loss amount (Rs. 25) and the cost price (Rs. 500):

$L\% = \left(\frac{25}{500}\right) \times 100$

$L\% = \frac{25}{5}$

$L\% = 5$

The loss percentage is 5%.

Therefore, if the article is sold for Rs. 475, there would be a loss of 5%.

Revision Table: Key Concepts in Profit and Loss

Concept	Definition	Formula
Cost Price (CP)	The price at which an article is purchased.	-
Selling Price (SP)	The price at which an article is sold.	-
Profit	When SP > CP.	Profit = SP - CP
Loss	When SP < CP.	Loss = CP - SP
Profit Percentage (P%)	Profit expressed as a percentage of CP.	$P\% = \left(\frac{\text{Profit}}{\text{CP}}\right) \times 100$
Loss Percentage (L%)	Loss expressed as a percentage of CP.	$L\% = \left(\frac{\text{Loss}}{\text{CP}}\right) \times 100$
SP with Profit		$SP = CP \times \left(1 + \frac{P\%}{100}\right)$
SP with Loss		$SP = CP \times \left(1 - \frac{L\%}{100}\right)$

Additional Information: Profit and Loss Problem Solving

Profit and loss problems are common in competitive exams and real-life financial calculations. Understanding the relationship between cost price, selling price, profit/loss amount, and profit/loss percentage is crucial.

- Always calculate profit or loss percentage on the **cost price** unless specified otherwise.
- A profit of X% means the selling price is $CP \times (1 + X/100)$.
- A loss of Y% means the selling price is $CP \times (1 - Y/100)$.
- These concepts are fundamental for more complex topics like discounts, marked price, and successive transactions.
- Practice different types of problems, including those where CP or SP is unknown, or where multiple transactions occur.

99. **Answer: d**

Explanation:

Understanding Statement and Assumptions Questions

In statement and assumptions questions, we are given a statement and then several assumptions. An assumption is something taken for granted or accepted as true without proof. We need to determine which of the given assumptions is 'implicit' in the statement, meaning it is directly implied or necessarily follows from the statement.

Analyzing the Given Statement

The statement provided is:

Statement: All the species of plants and animals are part of biodiversity and ecosystems and play major role in the overall health of the environment.

This statement emphasizes the critical importance of biodiversity (all plant and animal species) within ecosystems and its significant contribution to the overall health of the environment.

Analyzing the Given Assumptions

Let's evaluate each assumption based on the statement:

Assumption I Analysis

Assumption I: Preserve or create a backyard habitat. Save as many native plants as you can when building or landscaping.

- This assumption suggests specific actions people can take to help preserve biodiversity, such as creating backyard habitats and saving native plants.
- The statement tells us that biodiversity is important for environmental health. However, it does not mention *how* biodiversity should be preserved or what

specific actions are necessary.
- The statement focuses on the *role* and *importance* of biodiversity, not the *methods* of preservation.
- Therefore, Assumption I, which proposes specific actions, is not necessarily implied by the general statement about biodiversity's role.

Conclusion for Assumption I: Not implicit.

Assumption II Analysis

Assumption II: One of the major goals of sustainability is to preserve biodiversity. All life on Earth is connected through the flow of energy (planetary food web), and each time a species becomes endangered or lost to extinction, one more part of that energy flow is lost.

- This assumption connects preserving biodiversity to a major goal of sustainability. It also provides a reason why preserving biodiversity is important – the interconnectedness of life and energy flow, where species loss disrupts this flow.
- The statement says biodiversity plays a "major role in the overall health of the environment." If biodiversity is crucial for environmental health, it is highly logical that its preservation would be a significant goal, especially in contexts like sustainability which aims for long-term environmental well-being.
- The explanation about interconnectedness and energy flow supports the statement's claim that biodiversity plays a "major role." Losing parts (species) would indeed impact the "overall health" by disrupting these connections.
- Thus, the idea that preserving biodiversity is a goal linked to environmental health (like in sustainability) and the reason provided (interconnectedness) align well with the statement's core message about the importance of biodiversity for environmental health.

Conclusion for Assumption II: Implicit.

Final Conclusion

Based on the analysis, Assumption I proposes specific actions not implied by the statement's general assertion of importance. Assumption II, however, links biodiversity preservation to a goal (sustainability) and explains the underlying reason (interconnectedness), which directly supports and is implied by the statement about biodiversity's major role in environmental health.

Therefore, only Assumption II is implicit in the given statement.

Revision Table: Statement and Assumptions

Element	Content	Analysis	Implicit?
Statement	Biodiversity (plants and animals) are part of ecosystems and crucial for environment health.	Establishes the importance of biodiversity.	N/A
Assumption I	Preserve backyard habitat, save native plants.	Suggests specific preservation actions.	No. Statement doesn't specify *how* to preserve.
Assumption II	Preserving biodiversity is a sustainability goal; life is connected via energy flow; species loss breaks flow.	Connects preservation to a goal and explains *why* biodiversity is important (interconnectedness, energy flow, impact of loss).	Yes. Supports the statement's idea of biodiversity's major role in environmental health.

Additional Information: Biodiversity and Environmental Health

Biodiversity, which includes the variety of life forms from genes to ecosystems, is fundamental to the functioning of healthy ecosystems. Each species, no matter how small, can play a unique role in the complex web of life. For example, pollinators are essential for plant reproduction, decomposers break down organic matter, and predators control populations. The loss of any species can have ripple effects throughout an ecosystem, potentially leading to reduced resilience, decreased productivity, and instability. Preserving biodiversity is therefore not just about saving individual species but about maintaining the health and stability of the entire environment upon which all life, including humans, depends. Sustainability goals often focus on balancing human needs with the capacity of the environment to support life in the long term, making biodiversity preservation a natural fit.

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100. **Answer: d**

Explanation:

Solving Speed, Time, and Distance Problems

This problem involves the concepts of speed, time, and distance. We are given information about Madhu's travel to her university under two different speed conditions and her resulting late arrival times. The goal is to find the distance between her house and the university.

The fundamental relationship between speed, time, and distance is:

$$\text{Distance} = \text{Speed} \times \text{Time}$$

Let's define the variables:

- Let d be the distance from Madhu's house to the university (in km).
- Let t be the usual or scheduled time Madhu should take to reach the university (in hours).

The time taken can be expressed as:

$$\text{Time} = \frac{\text{Distance}}{\text{Speed}}$$

Analyzing Madhu's First Trip

In the first scenario, Madhu walks at a speed of 10 km/hr and is 15 minutes late.

- Speed (S_1) = 10 km/hr
- Time taken (T_1) = Usual time + Lateness

The lateness is 15 minutes. We need to convert this to hours:

$$15 \text{ minutes} = \frac{15}{60} \text{ hours} = \frac{1}{4} \text{ hours}$$

So, the time taken in the first case is $T_1 = t + \frac{1}{4}$ hours.

Using the distance formula:

$$d = S_1 \times T_1$$

$$d = 10 \times \left(t + \frac{1}{4}\right) \quad \text{(Equation 1)}$$

Analyzing Madhu's Second Trip

In the second scenario, Madhu increases her speed by 2 km/hr, so her new speed is $10 + 2 = 12$ km/hr. She is still 5 minutes late.

- Speed (S_2) = 12 km/hr
- Time taken (T_2) = Usual time + Lateness

The lateness is 5 minutes. Converting this to hours:

$$5 \text{ minutes} = \frac{5}{60} \text{ hours} = \frac{1}{12} \text{ hours}$$

So, the time taken in the second case is $T_2 = t + \frac{1}{12}$ hours.

Using the distance formula:

$$d = S_2 \times T_2$$

$$d = 12 \times \left(t + \frac{1}{12}\right) \quad \text{(Equation 2)}$$

Solving for the Distance

We have two equations for the same distance d. We can equate Equation 1 and Equation 2 to solve for the usual time t:

$$10 \times \left(t + \frac{1}{4}\right) = 12 \times \left(t + \frac{1}{12}\right)$$

Distribute the numbers on both sides:

$$10t + 10 \times \frac{1}{4} = 12t + 12 \times \frac{1}{12}$$

$$10t + \frac{10}{4} = 12t + 1$$

$$10t + 2.5 = 12t + 1$$

Now, let's rearrange the equation to solve for t. Subtract 10t from both sides:

$$2.5 = 12t - 10t + 1$$

$$2.5 = 2t + 1$$

Subtract 1 from both sides:

$$2.5 - 1 = 2t$$

$$1.5 = 2t$$

Divide by 2:

$$t = \frac{1.5}{2} = \frac{3/2}{2} = \frac{3}{4} \text{ hours}$$

The usual time Madhu should take is $\frac{3}{4}$ hours.

Calculating the Distance

Now that we have the value of t, we can substitute it into either Equation 1 or Equation 2 to find the distance d.

Using Equation 1:

$$d = 10 \times \left(t + \frac{1}{4}\right)$$

Substitute $t = \frac{3}{4}$:

$$d = 10 \times \left(\frac{3}{4} + \frac{1}{4}\right)$$

$$d = 10 \times \left(\frac{3+1}{4}\right)$$

$$d = 10 \times \left(\frac{4}{4}\right)$$

$$d = 10 \times 1$$

$$d = 10 \text{ km}$$

Let's verify using Equation 2:

$$d = 12 \times \left(t + \frac{1}{12}\right)$$

Substitute $t = \frac{3}{4}$:

$$d = 12 \times \left(\frac{3}{4} + \frac{1}{12}\right)$$

To add the fractions, find a common denominator, which is 12. $\frac{3}{4} = \frac{3 \times 3}{4 \times 3} = \frac{9}{12}$.

$$d = 12 \times \left(\frac{9}{12} + \frac{1}{12}\right)$$

$$d = 12 \times \left(\frac{9+1}{12}\right)$$

$$d = 12 \times \left(\frac{10}{12}\right)$$

$$d = 10 \text{ km}$$

Both equations give the same distance, 10 km.

Alternative Method: Using Time Difference

We can also solve this problem by focusing on the difference in time taken due to the change in speed.

- Let the distance be d km.

- Time taken at 10 km/hr = $\frac{d}{10}$ hours.
- Time taken at 12 km/hr = $\frac{d}{12}$ hours.

Madhu is 15 minutes late at 10 km/hr and 5 minutes late at 12 km/hr. The difference in her arrival times is $15 - 5 = 10$ minutes.

This 10-minute difference is the difference between the time taken at 10 km/hr and the time taken at 12 km/hr.

$$\text{Difference in time} = 10 \text{ minutes} = \frac{10}{60} \text{ hours} = \frac{1}{6} \text{ hours}$$

Since the speed is higher in the second case (12 km/hr), the time taken is less. So, the time taken at 10 km/hr is greater than the time taken at 12 km/hr by $\frac{1}{6}$ hours.

$$\frac{d}{10} - \frac{d}{12} = \frac{1}{6}$$

To solve for d, find a common denominator for 10, 12, and 6, which is 60. Multiply the entire equation by 60:

$$60 \times \left(\frac{d}{10} - \frac{d}{12}\right) = 60 \times \frac{1}{6}$$

$$\left(60 \times \frac{d}{10}\right) - \left(60 \times \frac{d}{12}\right) = 10$$

$$6d - 5d = 10$$

$$d = 10 \text{ km}$$

Both methods confirm that the distance is 10 km.

Summary of Results

Scenario	Speed (km/hr)	Lateness (minutes)	Time Taken (relative to usual time t)
1	10	15	$t + \frac{1}{4}$ hours
2	12	5	$t + \frac{1}{12}$ hours

By setting up equations based on the distance being constant, we found the usual time and subsequently the distance. The alternative method using the difference in time also yields the same result.

Conclusion

The distance of the university from Madhu's house is 10 km.

Revision Table: Speed, Time, Distance Concepts

Concept	Formula	Units (Common)	Notes
Speed	$\text{Speed} = \frac{\text{Distance}}{\text{Time}}$	km/hr, m/s	Rate of movement.
Distance	$\text{Distance} = \text{Speed} \times \text{Time}$	km, meters	Length covered during travel.
Time	$\text{Time} = \frac{\text{Distance}}{\text{Speed}}$	hours, seconds	Duration of travel.

It is crucial to maintain consistent units throughout the calculation. Convert minutes to hours or hours to minutes as needed.

Additional Information on Time and Distance Problems

Problems involving speed, time, and distance often appear in quantitative aptitude tests. They can involve various scenarios like:

- Constant speed travel.
- Changing speed over different parts of the journey.
- Problems involving late or early arrival times relative to a scheduled time.
- Problems involving relative speed (e.g., trains moving towards or away from each other).
- Problems involving boats and streams (upstream/downstream motion).

For problems with late or early arrivals, identifying the 'usual' or 'scheduled' time is key. The actual time taken is then expressed as $t \pm \text{difference}$.

- If late by x time, actual time = $t + x$.
- If early by y time, actual time = $t - y$.

In the problem solved above, since Madhu was late in both cases, the time taken was greater than the usual time t.

Understanding how a change in speed affects the time taken for a fixed distance is fundamental:

For a fixed distance, Speed is inversely proportional to Time. If speed increases, time decreases, and vice versa.

$$\text{If Distance (d) is constant, then } S \propto \frac{1}{T} \text{ or } S_1T_1 = S_2T_2 = d$$

This inverse relationship is implicitly used in the alternative method where the difference in times corresponds to the difference in times calculated using $\frac{d}{S}$.

Practice with various types of time and distance problems helps in quickly identifying the most efficient method for solving them.