

Algebraic Methods

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These notes were written to suit the contents of the course “Algebraic methods” given at NTU from August to October 2009 and 2010.

The main structure of the notes comes from the book by Robert Ash [1], a reference for this course.

The proof of Jordan-Holder Theorem was inspired by the one given by Stuart Rankin (www.math.uwo.ca/~srankin).

The presentation of the Primitive Element Theorem is based on the one given by Ken Brown (www.math.cornell.edu/~kbrown/4340/primitive.pdf).

Many parts in the Chapter on Galois Theory are based on the book by Stewart [6].

For the history comments, they are taken from [2, 3, 4, 5].

Finally, pictures are coming from Wikipedia.

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Chapter 1

Group Theory

Most lectures on group theory actually start with the definition of what is a group. It may be worth though spending a few lines to mention how mathematicians came up with such a concept.

Around 1770, Lagrange initiated the study of permutations in connection with the study of the solution of equations. He was interested in understanding solutions of polynomials in several variables, and got this idea to study the behaviour of polynomials when their roots are permuted. This led to what we now call Lagrange's Theorem, though it was stated as [5] *If a function $f(x_1, \dots, x_n)$ of n variables is acted on by all $n!$ possible permutations of the variables and these permuted functions take on only r values, then r is a divisor of $n!$* . It is Galois (1811-1832) who is considered by many as the founder of group theory. He was the first to use the term "group" in a technical sense, though to him it meant a collection of permutations closed under multiplication. Galois theory will be discussed much later in these notes. Galois was also motivated by the solvability of polynomial equations of degree n . From 1815 to 1844, Cauchy started to look at permutations as an autonomous subject, and introduced the concept of permutations generated by certain elements, as well as several notations still used today, such as the cyclic notation for permutations, the product of permutations, or the identity permutation. He proved what we call today Cauchy's Theorem, namely that if p is prime divisor of the cardinality of the group, then there exists a subgroup of cardinality p . In 1870, Jordan gathered all the applications of permutations he could find, from algebraic geometry, number theory, function theory, and gave a unified presentation (including the work of Cauchy and Galois). Jordan made explicit the notions of homomorphism, isomorphism (still for permutation groups), he introduced solvable groups, and proved that the indices in two composition series are the same (now called Jordan-Hölder Theorem). He also gave a proof that the alternating group A_n is simple for $n > 4$.

In 1870, while working on number theory (more precisely, in generalizing

Kummer's work on cyclotomic fields to arbitrary fields), Kronecker described in one of his papers a finite set of arbitrary elements on which he defined an abstract operation on them which satisfy certain laws, laws which now correspond to axioms for finite abelian groups. He used this definition to work with ideal classes. He also proved several results now known as theorems on abelian groups. Kronecker did not connect his definition with permutation groups, which was done in 1879 by Frobenius and Stickelberger.

Apart permutation groups and number theory, a third occurrence of group theory which is worth mentioning arose from geometry, and the work of Klein (we now use the term Klein group for one of the groups of order 4), and Lie, who studied transformation groups, that is transformations of geometric objects. The work by Lie is now a topic of study in itself, but Lie theory is beyond the scope of these notes.

The abstract point of view in group theory emerged slowly. It took something like one hundred years from Lagrange's work of 1770 for the abstract group concept to evolve. This was done by abstracting what was in common to permutation groups, abelian groups, transformation groups... In 1854, Cayley gave the modern definition of group for the first time:

"A set of symbols all of them different, and such that the product of any two of them (no matter in what order), or the product of any one of them into itself, belongs to the set, is said to be a group. These symbols are not in general convertible [commutative], but are associative."

Let us start from there.

1.1 Groups and subgroups

We start by introducing the object that will interest us for the whole chapter.

Definition 1.1. A **group** is a non-empty set G on which there is a binary operation $(a, b) \mapsto ab$ such that

- if a and b belong to G then ab is also in G (*closure*),
- $a(bc) = (ab)c$ for all a, b, c in G (*associativity*),
- there is an element $1 \in G$ such that $a1 = 1a = a$ for all $a \in G$ (*identity*),
- if $a \in G$, then there is an element $a^{-1} \in G$ such that $aa^{-1} = a^{-1}a = 1$ (*inverse*).

One can easily check that this implies the unicity of the identity and of the inverse.

A group G is called **abelian** if the binary operation is commutative, i.e., $ab = ba$ for all $a, b \in G$.

Remark. There are two standard notations for the binary group operation: either the additive notation, that is $(a, b) \mapsto a + b$ in which case the identity is

denoted by 0, or the multiplicative notation, that is $(a, b) \mapsto ab$ for which the identity is denoted by 1.

- Examples 1.1.**
1. \mathbb{Z} with the addition and 0 as identity is an abelian group.
 2. \mathbb{Z} with the multiplication is not a group since there are elements which are not invertible in \mathbb{Z} .
 3. The set of $n \times n$ invertible matrices with real coefficients is a group for the matrix product and identity the matrix \mathbf{I}_n . It is denoted by $GL_n(\mathbb{R})$ and called the **general linear group**. It is not abelian for $n \geq 2$.

The above examples are the easiest groups to think of. The theory of algebra however contains many examples of famous groups that one may discover, once equipped with more tools (for example, the Lie groups, the Brauer group, the Witt group, the Weyl group, the Picard group,...to name a few).

Definition 1.2. The **order** of a group G , denoted by $|G|$, is the cardinality of G , that is the number of elements in G .

We have only seen infinite groups so far. Let us look at some examples of finite groups.

- Examples 1.2.**
1. The **trivial group** $G = \{0\}$ may not be the most exciting group to look at, but still it is the only group of order 1.
 2. The group $G = \{0, 1, 2, \dots, n-1\}$ of integers modulo n is a group of order n . It is sometimes denoted by \mathbb{Z}_n (this should not be confused with p -adic integers though!).

Definition 1.3. A **subgroup** H of a group G is a non-empty subset of G that forms a group under the binary operation of G .

- Examples 1.3.**
1. If we consider the group $G = \mathbb{Z}_4 = \{0, 1, 2, 3\}$ of integers modulo 4, $H = \{0, 2\}$ is a subgroup of G .
 2. The set of $n \times n$ matrices with real coefficients and determinant of 1 is a subgroup of $GL_n(\mathbb{R})$, denoted by $SL_n(\mathbb{R})$ and called the **special linear group**.

At this point, in order to claim that the above examples are actually subgroups, one has to actually check the definition. The proposition below gives an easier criterion to decide whether a subset of a group G is actually a subgroup.

Proposition 1.1. *Let G be a group. Let H be a non-empty subset of G . The following are equivalent:*

1. H is a subgroup of G .
2. (a) $x, y \in H$ implies $xy \in H$ for all x, y .
(b) $x \in H$ implies $x^{-1} \in H$.

3. $x, y \in H$ implies $xy^{-1} \in H$ for all x, y .

Proof. We prove that $1. \Rightarrow 3. \Rightarrow 2. \Rightarrow 1.$

1. \Rightarrow 3. This part is clear from the definition of subgroup.
3. \Rightarrow 2. Since H is non-empty, let $x \in H$. By assumption of 3., we have that $xx^{-1} = 1 \in H$ and that $1x^{-1} \in H$ thus x is invertible in H . We now know that for $x, y \in H$, x and y^{-1} are in H , thus $x(y^{-1})^{-1} = xy$ is in H .
2. \Rightarrow 1. To prove this direction, we need to check the definition of group. Since closure and existence of an inverse are true by assumption of 2., and that associativity follows from the associativity in G , we are left with the existence of an identity. Now, if $x \in H$, then $x^{-1} \in H$ by assumption of 2., and thus $xx^{-1} = 1 \in H$ again by assumption of 2., which completes the proof. □

We will often use the last equivalence to check that a subset of a group G is a subgroup.

Now that we have these structures of groups and subgroups, let us introduce a map that allows to go from one group to another and that respects the respective group operations.

Definition 1.4. Given two groups G and H , a **group homomorphism** is a map $f : G \rightarrow H$ such that

$$f(xy) = f(x)f(y) \text{ for all } x, y \in G.$$

Note that this definition immediately implies that the identity 1_G of G is mapped to the identity 1_H of H . The same is true for the inverse, that is $f(x) = f(x)^{-1}$.

Example 1.4. The map $\exp : (\mathbb{R}, +) \rightarrow (\mathbb{R}^*, \cdot)$, $x \mapsto \exp(x)$ is a group homomorphism.

Definition 1.5. Two groups G and H are **isomorphic** if there is a group homomorphism $f : G \rightarrow H$ which is also a bijection.

Roughly speaking, isomorphic groups are “essentially the same”.

Example 1.5. If we consider again the group $G = \mathbb{Z}_4 = \{0, 1, 2, 3\}$ of integers modulo 4 with subgroup $H = \{0, 2\}$, we have that H is isomorphic to \mathbb{Z}_2 , the group of integers modulo 2.

A crucial definition is the definition of the order of a group element.

Definition 1.6. The **order** of an element $a \in G$ is the least positive integer n such that $a^n = 1$. If no such integer exists, the order of a is infinite. We denote it by $|a|$.

Note that the critical part of this definition is that the order is the *least* positive integer with the given property. The terminology *order* is used both for groups and group elements, but it is usually clear from the context which one is considered.

1.2 Cyclic groups

Let us now introduce a first family of groups, the cyclic groups.

Definition 1.7. A group G is **cyclic** if it is generated by a single element, which we denote by $G = \langle a \rangle$. We may denote by C_n a cyclic group of n elements.

Example 1.6. A finite cyclic group generated by a is necessarily abelian, and can be written (multiplicatively)

$$\{1, a, a^2, \dots, a^{n-1}\} \text{ with } a^n = 1$$

or (additively)

$$\{0, a, 2a, \dots, (n-1)a\} \text{ with } na = 0.$$

A finite cyclic group with n elements is isomorphic to the additive group \mathbb{Z}_n of integers modulo n .

Here are some properties of cyclic groups and its generators.

Proposition 1.2. *If G is a cyclic group of order n generated by a , the following conditions are equivalent:*

1. $|a^k| = n$.
2. k and n are relatively prime.
3. k has an inverse modulo n , that is there exists an integer s such that $ks \equiv 1$ modulo n .

Proof. Before starting the proof, recall that since a generates G of order n , we have that the order of a is n and in particular $a^n = 1$. The fact that $|a^k| = n$ means in words that the order of a^k is also n , that is, a^k is also a generator of G . We first prove that 1. \iff 2., while 2. \iff 3. follows from Bezout identity.

1. \implies 2. Suppose by contradiction that k and n are not relatively prime, that is, there exists $s > 1$ such that $s|k$ and $s|n$. Thus $n = ms$ and $k = sr$ for some $m, r \geq 1$ and we have

$$(a^k)^m = a^{srm} = a^{nr} = 1.$$

Now since $s > 1$, $m < n$, which contradicts that n is the order of a^k .

2. \implies 1. Suppose that the order of a^k is not n , then there exists $m < n$ such that $(a^k)^m = 1$ and thus $n|km$ since n is the order of a . If k and n were to be relatively prime, then n would divide m , which is a contradiction since $m < n$.

2. \Rightarrow 3. If k and n are relatively prime, then by Bezout identity, there exist r, s such that $1 = kr + ns$ and thus $kr \equiv 1$ modulo n .
3. \Rightarrow 2. If $kr \equiv 1$ modulo n then $1 = kr + ns$ for some s and the greatest common divisor of k and n must divide 1, which shows k and n are relatively prime.

□

Corollary 1.3. *The set of invertible elements modulo n forms a group under multiplication, whose order is the Euler function $\varphi(n)$, which by definition counts the number of positive integers less than n that are relatively prime to n .*

Example 1.7. Consider the group $\mathbb{Z}_6 = \{0, 1, 2, 3, 4, 5\}$, the group \mathbb{Z}_6^* of invertible elements in \mathbb{Z}_6 is $\mathbb{Z}_6^* = \{1, 5\}$. We have that $\varphi(6) = \varphi(2)\varphi(3) = 2$.

1.3 Cosets and Lagrange's Theorem

Definition 1.8. Let H be a subgroup of a group G . If $g \in G$, the **right coset** of H generated by g is

$$Hg = \{hg, h \in H\}$$

and similarly the **left coset** of H generated by g is

$$gH = \{gh, h \in H\}.$$

In additive notation, we get $H + g$ (which usually implies that we deal with a commutative group where we do not need to distinguish left and right cosets).

Example 1.8. If we consider the group $\mathbb{Z}_4 = \{0, 1, 2, 3\}$ and its subgroup $H = \{0, 2\}$ which is isomorphic to \mathbb{Z}_2 , the cosets of H in G are

$$0 + H = H, \quad 1 + H = \{1, 3\}, \quad 2 + H = H, \quad 3 + H = \{1, 3\}.$$

Clearly $0 + H = 2 + H$ and $1 + H = 3 + H$.

We see in the above example that while an element of $g \in G$ runs through all possible elements of the group G , some of the left cosets gH (or right cosets Hg) may be the same. It is easy to see when this exactly happens.

Lemma 1.4. *We have that $Ha = Hb$ if and only if $ab^{-1} \in H$ for $a, b \in G$. Similarly, $aH = bH$ if and only if $a^{-1}b \in H$ for $a, b \in G$.*

Proof. If two right cosets are the same, that is $Ha = Hb$, since H is a subgroup, we have $1 \in H$ and $a = hb$ for some $h \in H$, so $ab^{-1} = h \in H$.

Conversely, if $ab^{-1} = h \in H$, then $Ha = Hhb = Hb$, again since H is a subgroup. □

While one may be tempted to define a coset with a subset of G which is not a subgroup, we see that the above characterization really relies on the fact that H is actually a subgroup.

Example 1.9. It is thus no surprise that in the above example we have $0 + H = 2 + H$ and $1 + H = 3 + H$, since we have modulo 4 that $0 - 2 \equiv 2 \in H$ and $1 - 3 \equiv 2 \in H$.

Saying that two elements $a, b \in G$ generate the same coset is actually an **equivalence relation** in the following sense. We say that a is equivalent to b if and only if $ab^{-1} \in H$, and this relation satisfies the three properties of an equivalence relation:

- *reflexivity*: $aa^{-1} = 1 \in H$.
- *symmetry*: if $ab^{-1} \in H$ then $(ab^{-1})^{-1} = ba^{-1} \in H$.
- *transitivity*: if $ab^{-1} \in H$ and $bc^{-1} \in H$ then $(ab^{-1})(bc^{-1}) = ac^{-1} \in H$.

The **equivalence class** of a is the set of elements in G which are equivalent to a , namely

$$\{b, ab^{-1} \in H\}.$$

Since $ab^{-1} \in H \iff (ab^{-1})^{-1} = ba^{-1} \in H \iff b \in Ha$, we further have that

$$\{b, ab^{-1} \in H\} = Ha,$$

and a coset is actually an equivalence class.

Example 1.10. Let us get back to our example with the group $\mathbb{Z}_4 = \{0, 1, 2, 3\}$ and its subgroup $H = \{0, 2\}$. We compute the first coset $0 + H = H$, and thus we now know that the equivalence class of 0 is H , and thus there is no need to compute the coset generated by 2, since it will give the same coset. We then compute the coset $1 + H = \{1, 3\}$ and again there is no need to compute the one of 3 since it is already in the coset of 1. We thus get 2 cosets, and clearly they partition \mathbb{Z}_4 :

$$\mathbb{Z}_4 = \{0, 2\} \sqcup \{1, 3\} = H \sqcup (1 + H).$$

It is important to notice that the right (resp. left) cosets partition the group G (that the union of all cosets is G is clear since we run through all elements of G and H contains 1, and it is easy to see that if $x \in Ha$ and $x \in Hb$ then $Ha = Hb$).

Example 1.11. Consider \mathbb{R} as an additive group with subgroup \mathbb{Z} . Every real number up to addition by an integer looks like a number in $[0, 1)$. Thus

$$\mathbb{R} = \cup_{0 \leq x < 1} (x + \mathbb{Z}),$$

and the cosets of \mathbb{Z} partition \mathbb{R} .

Furthermore, since the map $h \mapsto ha$, $h \in H$, is a one-to-one correspondence, each coset has $|H|$ elements.

Definition 1.9. The **index** of a subgroup H in G is the number of right (left) cosets. It is a positive number or ∞ and is denoted by $[G : H]$.

If we think of a group G as being partitioned by cosets of a subgroup H , then the index of H tells how many times we have to translate H to cover the whole group.

Example 1.12. In Example 1.11, the index $[\mathbb{R} : \mathbb{Z}]$ is infinite, since there are infinitely many cosets of \mathbb{Z} in \mathbb{R} .

Theorem 1.5. (Lagrange's Theorem). *If H is a subgroup of G , then $|G| = |H|[G : H]$. In particular, if G is finite then $|H|$ divides $|G|$ and $[G : H] = |G|/|H|$.*

Proof. Let us start by recalling that the left cosets of H forms a partition of G , that is

$$G = \sqcup gH,$$

where g runs through a set of representatives (one for each coset). Let us look at the cardinality of G :

$$|G| = |\sqcup gH| = \sum |gH|$$

since we have a disjoint union of cosets, and the sum is again over the set of representatives. Now

$$\sum |gH| = \sum |H|$$

since we have already noted that each coset contains $|H|$ elements. We then conclude that

$$|G| = \sum |H| = [G : H]|H|.$$

□

Example 1.13. Consider $G = \mathbb{Z}$, $H = 3\mathbb{Z}$, then $[G : H] = 3$.

Of course, Lagrange did not prove Lagrange's theorem! The modern way of defining groups did not exist yet at his time. Lagrange was interested in polynomial equations, and in understanding the existence and nature of the roots (does every equation has a root? how many roots?...). What he actually proved was that if a polynomial in n variables has its variables permuted in all $n!$ ways, the number of different polynomials that are obtained is always a factor of $n!$. Since all the permutations of n elements are actually a group, the number of such polynomials is actually the index in the group of permutations of n elements of the subgroup H of permutations which preserve the polynomial. So the size of H divides $n!$, which is exactly the number of all permutations of n elements. This is indeed a particular case of what we call now Lagrange's Theorem.

Corollary 1.6. *1. Let G be a finite group. If $a \in G$, then $|a|$ divides $|G|$. In particular, $a^{|G|} = 1$.*

2. If G has prime order, then G is cyclic.

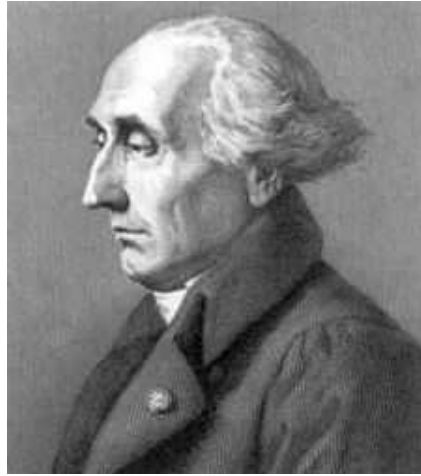


Figure 1.1: Joseph-Louis Lagrange (1736-1813)

Proof. 1. If $a \in G$ has order say m , then the subgroup $H = \{1, a, \dots, a^{m-1}\}$ is a cyclic subgroup of G with order $|H| = m$. Thus m divides $|G|$ by the theorem.

2. Since $|G|$ is prime, we may take $a \neq 1$ in G , and since the order of a has to divide $|G|$, we have $|a| = |G|$. Thus the cyclic group generated by a coincides with G .

□

Example 1.14. Using Lagrange's Theorem and its corollaries, we can already determine the groups of order from 1 to 5, up to isomorphism (see Table 1.1). If $|G|$ is prime, we now know that G is cyclic.

Let us look at the case where G is of order 4. Let $g \in G$. We know that the order of g is either 1, 2 or 4. If the order of g is 1, this is the identity. If G contains an element g of order 4, then that means that g generates the whole group, thus G is cyclic. If now G does not contain an element of order 4, then apart the identity, all the elements have order 2. From there, it is easy to obtain a multiplication table for G , and see that it coincides with the one of the group

$$\mathbb{Z}_2 \times \mathbb{Z}_2 = \{(x, y) \mid x, y \in \mathbb{Z}_2\}$$

with binary operation $(x, y) + (x', y') = (x + x', y + y')$. This group is called the **Klein group**, and it has several interpretations, for example, it is the group of isometries fixing a rectangle.

Remark. The above example also shows that the converse of Lagrange's Theorem is not true. If we take the group $G = C_2 \times C_2$, then 4 divides the order of G , however there is no element of order 4 in G .

$ G $	G
1	$\{1\}$
2	C_2
3	C_3
4	$C_4, C_2 \times C_2$
5	C_5

Table 1.1: Groups of order from 1 to 5. C_n denotes the cyclic group of order n

Once Lagrange's Theorem and its corollaries are proven, we can easily deduce Euler's and Fermat's Theorem.

Theorem 1.7. (Euler's Theorem). *If a and n are relatively prime positive integers, with $n \geq 2$, then*

$$a^{\varphi(n)} \equiv 1 \pmod{n}.$$

Proof. Since a and n are relatively prime, we know from Proposition 1.2 that a has an inverse modulo n , and by its corollary that the group of invertible elements has order $\varphi(n)$. Thus

$$a^{\varphi(n)} \equiv 1 \pmod{n}$$

by Lagrange's Theorem first corollary. □

Corollary 1.8. (Fermat's Little Theorem). *If p is a prime and a is a positive integer not divisible by p , then*

$$a^{p-1} \equiv 1 \pmod{p}.$$

This is particular case of Euler's Theorem when n is a prime, since then $\varphi(n) = p - 1$.

1.4 Normal subgroups and quotient group

Given a group G and a subgroup H , we have seen how to define the cosets of H , and thanks to Lagrange's Theorem, we already know that the number of cosets $[G : H]$ is related to the order of H and G by $|G| = |H|[G : H]$. A priori, the set of cosets of H has no structure. We are now interested in a criterion on H to give the set of its cosets a structure of group.

In what follows, we may write $H \leq G$ for H is a subgroup of G .

Definition 1.10. Let G be a group and $H \leq G$. We say that H is a **normal** subgroup of G , or that H is **normal** in G , if we have

$$cHc^{-1} = H, \text{ for all } c \in G.$$

We denote it $H \trianglelefteq G$, or $H \triangleleft G$ when we want to emphasize that H is a proper subgroup of G .

The condition for a subgroup to be normal can be stated in many slightly different ways.

Lemma 1.9. *Let $H \leq G$. The following are equivalent:*

1. $cHc^{-1} \subseteq H$ for all $c \in G$.
2. $cHc^{-1} = H$ for all $c \in G$, that is $cH = Hc$ for all $c \in G$.
3. Every left coset of H in G is also a right coset (and vice-versa, every right coset of H in G is also a left coset).

Proof. Clearly 2. implies 1., now $cHc^{-1} \subseteq H$ for all $c \in G$ if and only if $cH \subseteq Hc$. Let $x \in Hc$, that is $x = hc$ for some $h \in H$, so that

$$x = (cc^{-1})hc = c(c^{-1}hc) = ch'$$

for some $h' \in H$ since $cHc^{-1} \subseteq H$ for all c and thus in particular for c^{-1} . This shows that Hc is included in cH or equivalently that $H \subseteq cHc^{-1}$.

Also 2. clearly implies 3. Now suppose that $cH = Hd$. This means that c belongs to Hd by assumption and to Hc by definition, which means that $Hd = Hc$. \square

Example 1.15. Let $GL_n(\mathbb{R})$ be the group of $n \times n$ real invertible matrices, and let $SL_n(\mathbb{R})$ be the subgroup formed by matrices whose determinant is 1. Let us see that $SL_n(\mathbb{R}) \triangleleft GL_n(\mathbb{R})$.

For that, we have to check that $ABA^{-1} \in SL_n(\mathbb{R})$ for all $B \in SL_n(\mathbb{R})$ and $A \in GL_n(\mathbb{R})$. This is clearly true since

$$\det(ABA^{-1}) = \det(B) = 1.$$

Proposition 1.10. *If H is normal in G , then the cosets of H form a group.*

Proof. Let us first define a binary operation on the cosets: $(aH, bH) \mapsto (aH)(bH) = \{(ah)(bh')\}$, $ah \in aH$, $bh' \in bH$. We need to check that the definition of group is satisfied.

- **closure.** This is the part which asks a little bit of work. Since $cH = Hc$ for all $c \in G$, then

$$(aH)(bH) = a(Hb)H = a(bH)H = abHH = abH.$$

Note that this product does not depend on the choice of representatives.

- **Associativity** comes from G being associative.
- The **identity** is given by the coset $1H = H$.
- The **inverse** of the coset aH is $a^{-1}H$.

\square

Definition 1.11. The group of cosets of a normal subgroup N of G is called the **quotient group** of G by N . It is denoted by G/N .

Let us finish this section by discussing the connection between normal subgroups and homomorphisms. The first normal subgroup of interest will be the kernel of a group homomorphism.

Recall that if $f : G \rightarrow H$ is a group homomorphism, the **kernel** of f is defined by

$$\text{Ker}(f) = \{a \in G, f(a) = 1\}.$$

It is easy to see that $\text{Ker}(f)$ is a normal subgroup of G , since

$$f(aba^{-1}) = f(a)f(b)f(a)^{-1} = f(a)f(a)^{-1} = 1$$

for all $b \in \text{Ker}(f)$ and all $a \in G$.

The converse is more interesting.

Proposition 1.11. *Let G be a group. Every normal subgroup of G is the kernel of a homomorphism.*

Proof. Suppose that $N \trianglelefteq G$ and consider the map

$$\pi : G \rightarrow G/N, a \mapsto aN.$$

To prove the result, we have to show that π is a group homomorphism whose kernel is N . First note that π is indeed a map from group to group since G/N is a group by assuming that N is normal. Then we have that

$$\pi(ab) = abN = (aN)(bN) = \pi(a)\pi(b)$$

where the second equality comes from the group structure of G/N . Finally

$$\text{Ker}(\pi) = \{a \in G \mid \pi(a) = N\} = \{a \in G \mid aN = N\} = N.$$

□

Definition 1.12. Let $N \trianglelefteq G$. The group homomorphism

$$\pi : G \rightarrow G/N, a \mapsto aN$$

is called the **natural** or **canonical** map or projection.

Recall for further usage that for f a group homomorphism, we have the following characterization of injectivity: a homomorphism f is injective if and only if its kernel is trivial (that is, contains only the identity element). Indeed, if f is injective, then $\text{Ker}(f) = \{a, f(a) = 1\} = \{1\}$ since $f(1) = 1$. Conversely, if $\text{Ker}(f) = \{1\}$ and we assume that $f(a) = f(b)$, then

$$f(ab^{-1}) = f(a)f(b)^{-1} = f(a)f(a)^{-1} = 1$$

and $ab^{-1} = 1$ implying that $a = b$ and thus f is injective.

Terminology.

monomorphism=injective homomorphism

epimorphism=surjective homomorphism

isomorphism=bijective homomorphism

endomorphism=homomorphism of a group to itself

automorphism=isomorphism of a group with itself

We have looked so far at the particular subgroup of G which is its kernel. The proposition below describes more generally subgroups of G and H .

Proposition 1.12. *Let $f : G \rightarrow H$ be a homomorphism.*

1. *If K is a subgroup of G , then $f(K)$ is a subgroup of H . If f is an epimorphism and K is normal, then $f(K)$ is normal.*
2. *If K is a subgroup of H , then $f^{-1}(K) = \{x \in G, f(x) \in K\}$ is a subgroup of G . If K is normal, so is $f^{-1}(K)$.*

Proof. 1. To prove that $f(K)$ is a subgroup of H , it is enough to show that $f(a)f(b)^{-1} \in f(K)$ by Proposition 1.1, which is clear from

$$f(a)f(b)^{-1} = f(ab^{-1}) \in f(K).$$

If K is normal, we have to show that $cf(K)c^{-1} = f(K)$ for all $c \in H$. Since f is an epimorphism, there exists $d \in G$ such that $f(d) = c$, so that

$$cf(K)c^{-1} = f(d)f(K)f(d)^{-1} = f(dKd^{-1}) = f(K)$$

using that K is normal.

2. As before, to prove that $f^{-1}(K)$ is a subgroup of G , it is enough to show that $ab^{-1} \in f^{-1}(K)$ for $a, b \in f^{-1}(K)$, which is equivalent to show that $f(ab^{-1}) \in K$. This is now true since $f(ab^{-1}) = f(a)f(b)^{-1}$ with $a, b \in f^{-1}(K)$ and K a subgroup.

For the second claim, we have to show that

$$cf^{-1}(K)c^{-1} = f^{-1}(K) \iff f(cf^{-1}(K)c^{-1}) = K, \quad c \in G.$$

For $c \in G$ and $a \in f^{-1}(K)$, then

$$f(cac^{-1}) = f(c)f(a)f(c)^{-1} \in K$$

since K is normal. □

1.5 The isomorphism theorems

This section presents different isomorphism theorems which are important tools for proving further results. The first isomorphism theorem, that will be the second theorem to be proven after the factor theorem, is easier to motivate, since it will help us in computing quotient groups.

But let us first start with the so-called factor theorem. Assume that we have a group G which contains a normal subgroup N , another group H , and $f : G \rightarrow H$ a group homomorphism. Let π be the canonical projection (see Definition 1.12) from G to the quotient group G/N :

$$\begin{array}{ccc} G & \xrightarrow{f} & H \\ \pi \downarrow & \nearrow \bar{f} & \\ G/N & & \end{array}$$

We would like to find a homomorphism $\bar{f} : G/N \rightarrow H$ that makes the diagram commute, namely

$$f(a) = \bar{f}(\pi(a))$$

for all $a \in G$.

Theorem 1.13. (Factor Theorem). *Any homomorphism f whose kernel K contains N can be factored through G/N . In other words, there is a unique homomorphism $\bar{f} : G/N \rightarrow H$ such that $\bar{f} \circ \pi = f$. Furthermore*

1. \bar{f} is an epimorphism if and only if f is.
2. \bar{f} is a monomorphism if and only if $K = N$.
3. \bar{f} is an isomorphism if and only if f is an epimorphism and $K = N$.

Proof. Unicity. Let us start by proving that if there exists \bar{f} such that $\bar{f} \circ \pi = f$, then it is unique. Let \tilde{f} be another homomorphism such that $\tilde{f} \circ \pi = f$. We thus have that

$$(\bar{f} \circ \pi)(a) = (\tilde{f} \circ \pi)(a) = f(a)$$

for all $a \in G$, that is

$$\bar{f}(aN) = \tilde{f}(aN) = f(a).$$

This tells us that for all $bN \in G/N$ for which there exists an element b in G such that $\pi(b) = bN$, then its image by either \bar{f} or \tilde{f} is determined by $f(b)$. This shows that $\bar{f} = \tilde{f}$ by surjectivity of π .

Existence. Let $aN \in G/N$ such that $\pi(a) = aN$ for $a \in G$. We define

$$\bar{f}(aN) = f(a).$$

This is the most natural way to do it, however, we need to make sure that this is indeed well-defined, in the sense that it should not depend on the choice of

the representative taken in the coset. Let us thus take another representative, say $b \in aN$. Since a and b are in the same coset, they satisfy $a^{-1}b \in N \subset K$, where $K = \text{Ker}(f)$ by assumption. Since $a^{-1}b \in K$, we have $f(a^{-1}b) = 1$ and thus $f(a) = f(b)$.

Now that \bar{f} is well defined, let us check this is indeed a group homomorphism. First note that G/N is indeed a group since $N \trianglelefteq G$. Then, we have

$$\bar{f}(aNbN) = \bar{f}(abN) = f(ab) = f(a)f(b) = \bar{f}(aN)\bar{f}(bN)$$

and \bar{f} is a homomorphism.

1. The fact that \bar{f} is an epimorphism if and only if f is comes from the fact that both maps have the same image.
2. First note that the statement \bar{f} is a monomorphism if and only if $K = N$ makes sense since $K = \text{Ker}(f)$ is indeed a normal subgroup, as proved earlier.

To show that \bar{f} is a monomorphism is equivalent to show that $\text{Ker}(\bar{f})$ is trivial. By definition, we have

$$\begin{aligned} \text{Ker}(\bar{f}) &= \{aN \in G/N, \bar{f}(aN) = 1\} \\ &= \{aN \in G/N, \bar{f}(\pi(a)) = f(a) = 1\} \\ &= \{aN \in G/N, a \in K = \text{Ker}(f)\}. \end{aligned}$$

So the kernel of \bar{f} is exactly those cosets of the form aN with $a \in K$, but for the kernel to be trivial, we need it to be equal to N , that is we need $K = N$.

3. This is just a combination of the first two parts.

□

We are now ready to state the first isomorphism theorem.

Theorem 1.14. (1st Isomorphism Theorem). *If $f : G \rightarrow H$ is a homomorphism with kernel K , then the image of f is isomorphic to G/K :*

$$\text{Im}(f) \simeq G/\text{Ker}(f).$$

Proof. We know from the Factor Theorem that

$$\bar{f} : G/\text{Ker}(f) \rightarrow H$$

is an isomorphism if and only if f is an epimorphism, and clearly f is an epimorphism on its image, which concludes the proof. □

Example 1.16. We have seen in Example 1.15 that $SL_n(\mathbb{R}) \triangleleft GL_n(\mathbb{R})$. Consider the map

$$\det : GL_n(\mathbb{R}) \rightarrow (\mathbb{R}^*, \cdot),$$

which is a group homomorphism. We have that $\text{Ker}(\det) = SL_n(\mathbb{R})$. The 1st Isomorphism Theorem tells that

$$\text{Im}(\det) \simeq GL_n(\mathbb{R})/SL_n(\mathbb{R}).$$

It is clear that \det is surjective, since for all $a \in \mathbb{R}^*$, one can take the diagonal matrix with all entries at 1, but one which is a . Thus we conclude that

$$\mathbb{R}^* \simeq GL_n(\mathbb{R})/SL_n(\mathbb{R}).$$

The 1st Isomorphism Theorem can be nicely illustrated in terms of exact sequences.

Definition 1.13. Let F, G, H, I, \dots be groups, and let f, g, h, \dots be group homomorphisms. Consider the following sequence:

$$\cdots \quad F \xrightarrow{f} G \xrightarrow{g} H \xrightarrow{h} I \quad \cdots$$

We say that this sequence is **exact in one point** (say G) if $\text{Im}(f) = \text{Ker}(g)$. A sequence is **exact** if it is exact in all points.

A **short exact sequence** of groups is of the form

$$1 \xrightarrow{i} F \xrightarrow{f} G \xrightarrow{g} H \xrightarrow{j} 1$$

where i is the inclusion and j is the constant map 1.

Proposition 1.15. *Let*

$$1 \xrightarrow{i} F \xrightarrow{f} G \xrightarrow{g} H \xrightarrow{j} 1$$

be a short exact sequence of groups. Then $\text{Im}(f)$ is normal in G and we have a group isomorphism

$$G/\text{Im}(f) \simeq H,$$

or equivalently

$$G/\text{Ker}(g) \simeq H.$$

Proof. Since the sequence is exact, we have that $\text{Im}(f) = \text{Ker}(g)$ thus $\text{Im}(f)$ is a normal subgroup of G . By the first Isomorphism Theorem, we have that

$$G/\text{Ker}(g) \simeq \text{Im}(g) = H.$$

since $\text{Im}(g) = \text{Ker}(j) = H$. □

The formulation in terms of exact sequences is useful to know, since it happens very often in the literature that an exact sequence is given exactly to be able to compute such quotient groups.

Let us state the second and third isomorphism theorem.

Theorem 1.16. (2nd Isomorphism Theorem). *If H and N are subgroups of G , with N normal in G , then*

$$H/(H \cap N) \simeq HN/N.$$

There are many things to discuss about the statement of this theorem.

- First we need to check that HN is indeed a subgroup of G . To show that, notice that $HN = NH$ since N is a normal subgroup of G . This implies that for $hn \in HN$, its inverse $(hn)^{-1} = n^{-1}h^{-1} \in G$ actually lives in HN , and so does the product $(hn)(h'n') = h(nh')n'$.
- Note that by writing HN/N , we insist on the fact that there is no reason for N to be a subgroup of H . On the other hand, N is a normal subgroup of HN , since for all $hn \in HN$, we have

$$hnNn^{-1}h^{-1} = hNh^{-1} \subseteq N$$

since N is normal in G .

- We now know that the right hand side of the isomorphism is a quotient group. In order to see that so is the left hand side, we need to show that $H \cap N$ is a normal subgroup of H . This comes by noticing that $H \cap N$ is the kernel of the canonical map $\pi : G \rightarrow G/N$ restricted to H .

Now that all these remarks have been done, it is not difficult to see that the 2nd Isomorphism Theorem follows from the 1st Isomorphism Theorem, as does the 3rd Isomorphism Theorem.

Theorem 1.17. (3rd Isomorphism Theorem). *If N and H are normal subgroups of G , with N contained in H , then*

$$G/H \simeq (G/N)/(H/N).$$

1.6 Direct and semi-direct products

So far, we have seen how given a group G , we can get smaller groups, such as subgroups of G or quotient groups. We will now do the other way round, that is, starting with a collection of groups, we want to build larger new groups.

Let us start with two groups H and K , and let $G = H \times K$ be the cartesian product of H and K , that is

$$G = \{(h, k), h \in H, k \in K\}.$$

We define a binary operation on this set by doing componentwise multiplication (or addition if the binary operations of H and K are denoted additively) on G :

$$(h_1, k_1)(h_2, k_2) = (h_1h_2, k_1k_2) \in H \times K.$$

Clearly G is closed under multiplication, its operation is associative (since both operations on H and K are), it has an identity element given by $(1_H, 1_K)$ and the inverse of (h, k) is (h^{-1}, k^{-1}) . In summary, G is a group.

Definition 1.14. Let H, K be two groups. The group $G = H \times K$ with binary operation defined componentwise as described above is called the **external direct product** of H and K .

Examples 1.17. 1. The Klein group $C_2 \times C_2$ of order 4 is a direct product (here C_2 denotes the cyclic group of order 2). It can also be written $\mathbb{Z}_2 \times \mathbb{Z}_2$ with an additive law.

2. The group $(\mathbb{R}, +) \times (\mathbb{R}, +)$ with componentwise addition is a direct product.

Note that G contains isomorphic copies \bar{H} and \bar{K} of respectively H and K , given by

$$\bar{H} = \{(h, 1_K), h \in H\}, \quad \bar{K} = \{(1_H, k), k \in K\},$$

which furthermore are normal subgroups of G . Let us for example see that \bar{H} is normal in G . By definition, we need to check that

$$(h, k)\bar{H}(h^{-1}, k^{-1}) \subseteq \bar{H}, \quad (h, k) \in G.$$

Let $(h', 1_K) \in \bar{H}$, we compute that

$$(h, k)(h', 1_K)(h^{-1}, k^{-1}) = (hh'h^{-1}, 1_K) \in \bar{H},$$

since $hh'h^{-1} \in H$. The same computation holds for \bar{K} .

If we gather what we know about G, \bar{H} and \bar{K} , we get that

- by definition, $G = \bar{H}\bar{K}$ and $\bar{H} \cap \bar{K} = \{1\}$,
- by what we have just proved, \bar{H} and \bar{K} are two normal subgroups of G .

This motivates the following definition.

Definition 1.15. If a group G contains normal subgroups H and K such that $G = HK$ and $H \cap K = \{1\}$, we say that G is the **internal direct product** of H and K .

The next result explicits the connection between internal and external products.

Proposition 1.18. *If G is the internal direct product of H and K , then G is isomorphic to the external direct product $H \times K$.*

Proof. To show that G is isomorphic to $H \times K$, we define the following map

$$f : H \times K \rightarrow G, \quad f(h, k) = hk.$$

First remark that if $h \in H$ and $k \in K$, then $hk = kh$. Indeed, we have using that both K and H are normal that

$$(hkh^{-1})k^{-1} \in K, \quad h(kh^{-1}k^{-1}) \in H$$

implying that

$$hkh^{-1}k^{-1} \in K \cap H = \{1\}.$$

We are now ready to prove that f is a group isomorphism.

1. This is a group homomorphism since

$$f((h, k)(h', k')) = f(hh', kk') = h(h'k)k' = h(kh')k' = f(h, k)f(h', k').$$

2. The map f is injective. This can be seen by checking that its kernel is trivial. Indeed, if $f(h, k) = 1$ then

$$hk = 1 \Rightarrow h = k^{-1} \Rightarrow h \in K \Rightarrow h \in H \cap K = \{1\}.$$

We have then that $h = k = 1$ which proves that the kernel is $(1, 1)$.

3. The map f is surjective since by definition $G = HK$.

□

Note that the definitions of external and internal product are surely not restricted to two groups. One can in general define them for n groups H_1, \dots, H_n . Namely

Definition 1.16. If H_1, \dots, H_n are arbitrary groups, the **external direct product** of the H_i is the cartesian product

$$G = H_1 \times H_2 \times \cdots \times H_n$$

with componentwise multiplication.

If G contains normal subgroups H_1, \dots, H_n such that $G = H_1 \cdots H_n$ and each g can be represented as $h_1 \cdots h_n$ uniquely, we say that G is the **internal direct product** of the H_i .

We can see a slight difference in the definition of internal product, since in the case of two subgroups, the condition given was not that each g can be represented uniquely as $h_1 h_2$, but instead that the intersection of the two subgroups is $\{1\}$. The next proposition shows the connection between these two points of view.

Proposition 1.19. *Suppose that $G = H_1 \cdots H_n$ where each H_i is a normal subgroup of G . The following conditions are equivalent.*

1. G is the internal direct product of the H_i .
2. $H_1 H_2 \cdots H_{i-1} \cap H_i = \{1\}$, for all $i = 1, \dots, n$.

Proof. Let us prove 1. \iff 2.

1. \Rightarrow 2. Let us assume that G is the internal direct product of the H_i , which means that every element in G can be written uniquely as a product of elements in H_i . Now let us take $g \in H_1 H_2 \cdots H_{i-1} \cap H_i = \{1\}$. We have that $g \in H_1 H_2 \cdots H_{i-1}$, which is uniquely written as $g = h_1 h_2 \cdots h_{i-1} 1_{H_i} \cdots 1_{H_n}$, $h_j \in H_j$. On the other hand, $g \in H_i$ thus $g = 1_{H_1} \cdots 1_{H_{i-1}} g$ and by unicity of the representation, we have $h_j = 1$ for all j and $g = \{1\}$.

2. \Rightarrow 1. Conversely, let us assume that $g \in G$ can be written either

$$g = h_1 h_2 \cdots h_n, \quad h_j \in H_j,$$

or

$$g = k_1 k_2 \cdots k_n, \quad k_j \in H_j.$$

Recall that since all H_j are normal subgroups, then

$$h_i h_j = h_j h_i, \quad h_i \in H_i, \quad h_j \in H_j.$$

(If you cannot recall the argument, check out the proof of Proposition 1.18). This means that we can do the following manipulations:

$$\begin{aligned} h_1 h_2 \cdots h_n &= k_1 k_2 \cdots k_n \\ \Leftrightarrow h_2 \cdots h_n &= \underbrace{(h_1^{-1} k_1)}_{\in H_1} k_2 \cdots k_n \\ \Leftrightarrow h_3 \cdots h_n &= (h_1^{-1} k_1)(h_2^{-1} k_2) k_3 \cdots k_n \end{aligned}$$

and so on and so forth till we reach

$$h_n k_n^{-1} = (h_1^{-1} k_1) \cdots (h_{n-1}^{-1} k_{n-1}).$$

Since the left hand side belongs to H_n while the right hand side belongs to $H_1 \cdots H_{n-1}$, we get that

$$h_n k_n^{-1} \in H_n \cap H_1 \cdots H_{n-1} = \{1\},$$

implying that $h_n = k_n$. We conclude the proof by iterating this process. □

Let us get back to the case of two groups. We have seen above that we can endow the cartesian product of two groups H and K with a group structure by considering componentwise binary operation

$$(h_1, k_1)(h_2, k_2) = (h_1 h_2, k_1 k_2) \in H \times K.$$

The choice of this binary operation of course determines the structure of $G = H \times K$, and in particular we have seen that the isomorphic copies of H and K in G are normal subgroups. Conversely in order to define an internal direct product, we need to assume that we have two normal subgroups.

We now consider a more general setting, where the subgroup K does not have to be normal (and will not be in general), for which we need to define a new binary operation on the cartesian product $H \times K$. This will lead us to the definition of internal and external semi-direct product.

Recall that an automorphism of a group H is a bijective group homomorphism from H to H . It is easy to see that the set of automorphisms of H forms a group with respect to the composition of maps and identity element the identity map Id_H . We denote it by $\text{Aut}(H)$.

Proposition 1.20. *Let H and K be groups, and let*

$$\rho : K \rightarrow \text{Aut}(H), \quad k \mapsto \rho_k$$

be a group homomorphism. Then the binary operation

$$(H \times K) \times (H \times K) \rightarrow (H \times K), \quad ((h, k), (h', k')) \mapsto (h\rho_k(h'), kk')$$

endows $H \times K$ with a group structure, with identity element $(1, 1)$.

Proof. First notice that the closure property is satisfied.

(Identity). Let us show that $(1, 1)$ is the identity element. We have

$$(h, k)(1, 1) = (h\rho_k(1), k) = (h, k)$$

for all $h \in H, k \in K$, since ρ_k is a group homomorphism. We also have

$$(1, 1)(h', k') = (\rho_1(h'), k') = (h', k')$$

for all $h' \in H, k' \in K$, since ρ being a group homomorphism, it maps 1_K to $1_{\text{Aut}(K)} = \text{Id}_H$.

(Inverse). Let $(h, k) \in H \times K$ and let us show that $(\rho_k^{-1}(h^{-1}), k^{-1})$ is the inverse of (h, k) . We have

$$(h, k)(\rho_k^{-1}(h^{-1}), k^{-1}) = (h\rho_k(\rho_k^{-1}(h^{-1})), 1) = (hh^{-1}, 1) = (1, 1).$$

We also have

$$\begin{aligned} (\rho_k^{-1}(h^{-1}), k^{-1})(h, k) &= (\rho_k^{-1}(h^{-1})\rho_{k^{-1}}(h), 1) \\ &= (\rho_{k^{-1}}(h^{-1})\rho_{k^{-1}}(h), 1) \end{aligned}$$

using that $\rho_k^{-1} = \rho_{k^{-1}}$ since ρ is a group homomorphism. Now

$$(\rho_{k^{-1}}(h^{-1})\rho_{k^{-1}}(h), 1) = (\rho_{k^{-1}}(h^{-1}h), 1) = (\rho_{k^{-1}}(1), 1) = (1, 1)$$

using that $\rho_{k^{-1}}$ is a group homomorphism for all $k \in K$.

Associativity. This is the last thing to check. On the one hand, we have

$$\begin{aligned} [(h, k)(h', k')](h'', k'') &= (h\rho_k(h'), kk')(h'', k'') \\ &= (h\rho_k(h')\rho_{kk'}(h''), (kk')k''), \end{aligned}$$

while on the other hand

$$\begin{aligned} (h, k)[(h', k')(h'', k'')] &= (h, k)(h'\rho_{k'}(h''), k'k'') \\ &= (h\rho_k(h'\rho_{k'}(h'')), k(k'k'')). \end{aligned}$$

Since K is a group, we have $(kk')k'' = k(k'k'')$. We now look at the first component. Note that $\rho_{kk'} = \rho_k \circ \rho_{k'}$ using that ρ is a group homomorphism, so that

$$h\rho_k(h')\rho_{kk'}(h'') = h\rho_k(h')\rho_k(\rho_{k'}(h'')).$$

Furthermore, ρ_k is a group homomorphism, yielding

$$h\rho_k(h')\rho_k(\rho_{k'}(h'')) = h\rho_k(h'\rho_{k'}(h''))$$

which concludes the proof. \square

We are now ready to define the first semi-direct product.

Definition 1.17. Let H and K be two groups, and let

$$\rho : K \rightarrow \text{Aut}(H)$$

be a group homomorphism. The set $H \times K$ endowed with the binary operation

$$((h, k), (h', k')) \mapsto (h\rho_k(h'), kk')$$

is called an **external semi-direct product** of H and K by ρ , denoted by $H \times_\rho K$.

We can make observations similar to what we did for direct products. Namely, we can identify two isomorphic copies \bar{H} and \bar{K} of respectively H and K , given by

$$\bar{H} = \{(h, 1_K), h \in H\}, \quad \bar{K} = \{(1_H, k), k \in K\},$$

and look at the properties of these subgroups.

- The subgroup $\bar{H} = \{(h, 1), h \in H\}$ is normal in $H \times_\rho K$, this can be seen by writing down the definition of normal subgroup. (We cannot claim the same for \bar{K} !).
- We have $\bar{H}\bar{K} = H \times_\rho K$, since every element $(h, k) \in H \times_\rho K$ can be written as $(h, 1)(1, k)$ (indeed $(h, 1)(1, k) = (h\rho_1(1), k) = (h, k)$).
- We have $\bar{H} \cap \bar{K} = \{1\}$.

This motivates the definition of internal semi-direct products.

Definition 1.18. Let G be a group with subgroups H and K . We say that G is the **internal semi-direct product** of H and K if H is a normal subgroup of G , such that $HK = G$ and $H \cap K = \{1\}$. It is denoted by

$$G = H \rtimes K.$$

Example 1.18. The dihedral group D_n is the group of all reflections and rotations of a regular polygon with n vertices centered at the origin. It has order $2n$. Let a be a rotation of angle $2\pi/n$ and let b be a reflection. We have that

$$D_n = \{a^i b^j, 0 \leq i \leq n-1, j = 0, 1\},$$

with

$$a^n = b^2 = (ba)^2 = 1.$$

We thus have that $\langle a \rangle = C_n$ and $\langle b \rangle = C_2$, where C_n denotes the cyclic group n . To prove that

$$D_n \simeq C_n \rtimes C_2,$$

we are left to check that $\langle a \rangle \cap \langle b \rangle = \{1\}$ and that $\langle a \rangle$ is normal in D_n . The former can be seen geometrically (a reflection cannot be obtained by rotations) while the latter reduces to show that

$$bab^{-1} \in \langle a \rangle,$$

which can be checked using $b^2 = (ba)^2 = 1$.

Again similarly to the case of direct products, these assumptions guarantee that we can write uniquely elements of the internal semi-direct product. Let us repeat things explicitly.

Lemma 1.21. *Let G be a group with subgroups H and K . Suppose that $G = HK$ and $H \cap K = \{1\}$. Then every element g of G can be written uniquely in the form hk , for $h \in H$ and $k \in K$.*

Proof. Since $G = HK$, we know that g can be written as hk . Suppose it can also be written as $h'k'$. Then $hk = h'k'$ so $h'^{-1}h = k'k^{-1} \in H \cap K = \{1\}$. Therefore $h = h'$ and $k = k'$. \square

The internal and external direct products were two sides of the same objects, so are the internal and external semi-direct products. If $G = H \times_{\rho} K$ is the external semi-direct product of H and K , then $\bar{H} = H \times \{1\}$ is a normal subgroup of G and it is clear that G is the internal semi-direct product of $H \times \{1\}$ and $\{1\} \times K$. This reasoning allows us to go from external to internal semi-direct products. The result below goes in the other direction, from internal to external semi-direct products.

Proposition 1.22. *Suppose that G is a group with subgroups H and K , and G is the internal semi-direct product of H and K . Then $G \simeq H \times_{\rho} K$ where $\rho : K \rightarrow \text{Aut}(H)$ is given by $\rho_k(h) = khk^{-1}$, $k \in K$, $h \in H$.*

Proof. Note that ρ_k belongs to $\text{Aut}(H)$ since H is normal.

By the above lemma, every element g of G can be written uniquely in the form hk , with $h \in H$ and $k \in K$. Therefore, the map

$$\varphi : H \times_{\rho} K \rightarrow G, \quad \varphi(h, k) = hk$$

is a bijection. It only remains to show that this bijection is a homomorphism.

Given (h, k) and (h', k') in $H \times_{\rho} K$, we have

$$\varphi((h, k)(h', k')) = \varphi((h\rho_k(h'), kk')) = \varphi(hkh'k^{-1}, kk') = hkh'k' = \varphi(h, k)\varphi(h', k').$$

Therefore φ is a group isomorphism, which concludes the proof. \square

In words, we have that every internal semi-direct product is isomorphic to some external semi-direct product, where ρ is the conjugation.

Example 1.19. Consider the dihedral group D_n from the previous example:

$$D_n \simeq C_n \rtimes C_2.$$

According to the above proposition, D_n is isomorphic to an external semi-direct product

$$D_n \simeq C_n \times_{\rho} C_2,$$

where

$$\rho : C_2 \rightarrow \text{Aut}(C_n),$$

maps to the conjugation in $\text{Aut}(C_n)$. Since C_n is abelian, we have explicitly that

$$1 \mapsto \rho_1 = \text{Id}_{C_n}, \quad b \mapsto \rho_b, \quad \rho_b(a) = a^{-1}.$$

$ G $	G abelian	G non-abelian
1	$\{1\}$	-
2	C_2	-
3	C_3	-
4	$C_4, C_2 \times C_2$	-
5	C_5	-
6	$C_6 = C_3 \times C_2$	$D_3 = C_3 \rtimes C_2$
7	C_7	-
8	$C_8, C_4 \times C_2, C_2 \times C_2 \times C_2$	$D_4 = C_4 \rtimes C_2$

Table 1.2: C_n denotes the cyclic group of order n , D_n the dihedral group

Before finishing this section, note the following distinction: the external (semi-)direct product of groups allows to construct new groups starting from different abstract groups, while the internal (semi-)direct product helps in analyzing the structure of a given group.

Example 1.20. Thanks to the new structures we have seen in this section, we can go on our investigation of groups of small orders. We can get two new groups of order 6 and 4 of order 8:

- $C_3 \times C_2$ is the direct product of C_3 and C_2 . You may want to check that it is actually isomorphic to C_6 .
- The dihedral group $D_3 = C_3 \rtimes C_2$ is the semi-direct product of C_3 and C_2 . We get similarly $D_4 = C_4 \rtimes C_2$.
- The direct product $C_4 \times C_2$ and the direct product of the Klein group $C_2 \times C_2$ with C_2 .

The table actually gives an exact classification of groups of small order (except the missing non-abelian quaternion group of order 8), though we have not proven it.

1.7 Group action

Since we introduced the definition of group as a set with a binary operation which is closed, we have been computing things internally, that is inside a group structure. This was the case even when considering cartesian products of groups, where the first thing we did was to endow this set with a group structure.

In this section, we wonder what happens if we have a group and a set, which may or may not have a group structure. We will define a group action, that is a way to do computations with two objects, one with a group law, not the other one.

As a first result, we will prove the so-called Cayley's theorem, whose proof will motivate the introduction of group action. Since the statement of this theorem uses permutation groups, we start by recalling the necessary definitions.

Definition 1.19. A **permutation** of a set S is a bijection on S . The set of all such functions (with respect to function composition) is a group called the **symmetric group** on S . We denote by S_n the symmetric group on n elements.

Example 1.21. Consider the symmetric group S_3 of permutations on 3 elements. It is given by

$$\begin{aligned} e &: 123 \rightarrow 123 \text{ or } () \\ a &: 123 \rightarrow 213 \text{ or } (12) \\ b &: 123 \rightarrow 132 \text{ or } (23) \\ ab &: 123 \rightarrow 312 \text{ or } (132) \\ ba &: 123 \rightarrow 231 \text{ or } (123) \\ aba &: 123 \rightarrow 321 \text{ or } (13) \end{aligned}$$

One can check that this is indeed a group. The notation (132) means that the permutation sends 1 to 3, 3 to 2, and 2 to 1.

Theorem 1.23. (Cayley's Theorem.) *Every group is isomorphic to a group of permutations.*

Proof. Let S_G be the group of permutations of G . We will prove that every group is isomorphic to a subgroup of S_G . The idea is that each element $g \in G$ corresponds to a permutation of G , namely we need to define a map from G to S_G :

$$\lambda : G \rightarrow S_G, g \mapsto \lambda(g) = \lambda_g$$

and since λ_g is a bijection on G , we need to define what λ_g does:

$$\lambda_g : G \rightarrow G, \lambda_g(x) = gx.$$

For justifying that λ_g is indeed a bijection, it is enough to see that g^{-1} exists since G is a group (try to write down the definition of injection and surjection).

We are left to check that λ is an injective group homomorphism. Injectivity again comes from G being a group, for if $\lambda_g(x) = \lambda_h(x)$ for all $x \in G$, then it has to be true that $gx = hx$ when $x = 1$.

Now

$$(\lambda(a) \circ \lambda(b))(x) = (\lambda_a \circ \lambda_b)(x) = a(bx) = \lambda_{ab}(x) = \lambda(ab)(x)$$

for all x , so that $\lambda(a) \circ \lambda(b) = \lambda(ab)$ which concludes the proof. \square

Examples 1.22. 1. Consider the group $\{0, 1\}$ of integers modulo 2. The group element 0 corresponds to the identity permutation, while the group element 1 corresponds to the permutation (12).

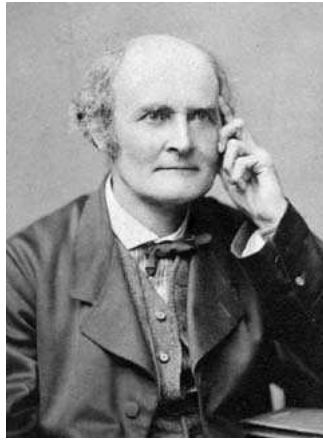


Figure 1.2: Arthur Cayley (1821-1895): he was the first to define the concept of a group in the modern way. Before him, groups referred to permutation groups.

2. Let us consider the group $\{0, 1, 2\}$ of integers modulo 3 to get a less simple example. Again 0 corresponds to the identity permutation, 1 corresponds to the permutation (123), and 2 to the permutation (132). To see that it makes sense, you may want to check that the arithmetic works similarly on both sides. For example, we can say that $1 + 1 = 2$ on the one hand, now on the other hand, we have $(123)(123) = (132)$.
3. One can check that the dihedral group D_3 of order 6 is isomorphic to S_3 (this can be done for example by working out the multiplication table for each group).

The key point in the proof of Cayley's Theorem is the way the function λ_g is defined. We see that for $x \in G$, g "acts" (via λ_g) on x by multiplication.

Definition 1.20. The group G *acts* on the set X if for all $g \in G$, there is a map

$$G \times X \rightarrow X, (g, x) \mapsto g \cdot x$$

such that

1. $h \cdot (g \cdot x) = (hg) \cdot x$ for all $g, h \in G$, for all $x \in X$.
2. $1 \cdot x = x$ for all $x \in X$.

The first condition says that we have two laws, the group law between elements of the group, and the action of the group on the set, which are compatible.

Examples 1.23. Let us consider two examples where a group G acts on itself.

1. Every group acts on itself by left multiplication. This is called the regular action.

2. Every group acts on itself by conjugation. Let us write this action as

$$g \cdot x = gxg^{-1}.$$

Let us check the action is actually well defined. First, we have that

$$h \cdot (g \cdot x) = h \cdot (gxg^{-1}) = hgxg^{-1}h^{-1} = (hg)xg^{-1}h^{-1} = (hg) \cdot x.$$

As for the identity, we get

$$1 \cdot x = 1x1^{-1} = x.$$

Similarly to the notion of kernel for a homomorphism, we can define the kernel of an action.

Definition 1.21. The **kernel** of an action $G \times X \rightarrow X$, $(g, x) \mapsto g \cdot x$ is given by

$$\text{Ker} = \{g \in G, g \cdot x = x \text{ for all } x\}.$$

This is the set of elements of G that fix everything in X . When the group G acts on itself, that is $X = G$ and the action is the conjugation, we have

$$\text{Ker} = \{g \in G, gxg^{-1} = x \text{ for all } x\} = \{g \in G, gx = xg \text{ for all } x\}.$$

This is called the **center** of G , denoted by $Z(G)$.

Definition 1.22. Suppose that a group G acts on a set X . The **orbit** $B(x)$ of x under the action of G is defined by

$$B(x) = \{g \cdot x, g \in G\}.$$

This means that we fix an element $x \in X$, and then we let g act on x when g runs through all the elements of G . By the definition of an action, $g \cdot x$ belongs to X , so the orbit gives a subset of X .

It is important to notice that orbits partition X . Clearly, one has that $X = \cup_{x \in X} B(x)$. But now, assume that one element x of X belongs to two orbits $B(y)$ and $B(z)$, then it means that $x = g \cdot y = g' \cdot z$, which in turn implies, due to the fact that G is a group, that

$$y = g^{-1}g' \cdot z, z = (g')^{-1}g \cdot y.$$

In words, that means that y belongs to the orbit of z , and vice-versa, z belongs to the orbit of y , and thus $B(y) = B(z)$. We can then pick a set of representatives for each orbit, and write that

$$X = \sqcup B(x),$$

where the disjoint union is taken over the set of representatives.

Definition 1.23. Suppose that a group G acts on a set X . We say that the action is **transitive**, or that G **acts transitively** on X if there is only one orbit, namely, for all $x, y \in X$, there exists $g \in G$ such that $g \cdot x = y$.

Definition 1.24. The **stabilizer** of an element $x \in X$ under the action of G is defined by

$$\text{Stab}(x) = \{g \in G, g \cdot x = x\}.$$

Given x , the stabilizer $\text{Stab}(x)$ is the set of elements of G that leave x fixed. One may check that this is a subgroup of G . We have to check that if $g, h \in \text{Stab}(x)$, then $gh^{-1} \in \text{Stab}(x)$. Now

$$(gh^{-1}) \cdot x = g \cdot (h^{-1} \cdot x)$$

by definition of action. Since $h \in \text{Stab}(x)$, we have $h \cdot x = x$ or equivalently $x = h^{-1} \cdot x$, so that

$$g \cdot (h^{-1} \cdot x) = g \cdot x = x,$$

which shows that $\text{Stab}(x)$ is a subgroup of G .

Examples 1.24. 1. The regular action (see the previous example) is transitive, and for all $x \in X = G$, we have $\text{Stab}(x) = \{1\}$, since x is invertible and we can multiply $g \cdot x = x$ by x^{-1} .

2. Let us consider the action by conjugation, which is again an action of G on itself ($X = G$): $g \cdot x = gxg^{-1}$. The action has no reason to be transitive in general, and for all $x \in X = G$, the orbit of x is given by

$$B(x) = \{gxg^{-1}, g \in G\}.$$

This is called the **conjugacy class** of x . Let us now consider the stabilizer of an element $x \in X$:

$$\text{Stab}(x) = \{g \in G, gxg^{-1} = x\} = \{g \in G, gx = xg\},$$

which is the **centralizer** of x , that we denote by $C_G(x)$.

Note that we can define similarly the centralizer $C_G(S)$ where S is an arbitrary subset of G as the set of elements of G which commute with everything in S . The two extreme cases are: if $S = \{x\}$, we get the centralizer of one element, if $S = G$, we get the center $Z(G)$.

Theorem 1.24. (The Orbit-Stabilizer Theorem). *Suppose that a group G acts on a set X . Let $B(x)$ be the orbit of $x \in X$, and let $\text{Stab}(x)$ be the stabilizer of x . Then the size of the orbit is the index of the stabilizer, that is*

$$|B(x)| = [G : \text{Stab}(x)].$$

If G is finite, then

$$|B(x)| = |G|/|\text{Stab}(x)|.$$

In particular, the size of the orbit divides the order of the group.

Proof. Recall first that $[G : \text{Stab}(x)]$ counts the number of left cosets of $\text{Stab}(x)$ in G , that is the cardinality of

$$G/\text{Stab}(x) = \{g\text{Stab}(x), g \in G\}.$$

Note that cosets of $\text{Stab}(x)$ are well-defined since we saw that $\text{Stab}(x)$ is a subgroup of G . The idea of the proof is to build a function between the sets $B(x)$ and $G/\text{Stab}(x)$ which is a bijection. That the cardinalities are the same will then follow.

Take $y \in B(x)$, that is $y = g \cdot x$ for some $g \in G$. We define a map

$$f : B(x) \rightarrow G/\text{Stab}(x), y = g \cdot x \mapsto g\text{Stab}(x).$$

Before checking that this map is a bijection, we need to check that it is well defined. Indeed, for a given y , there is no reason for the choice of g to be unique (there is in general no bijection between G and $B(x)$). Suppose that

$$y = g_1 \cdot x = g_2 \cdot x$$

then

$$g_2^{-1}g_1 \cdot x = x \iff g_1\text{Stab}(x) = g_2\text{Stab}(x).$$

The equivalence is the characterization of having two equal cosets. This is exactly what we wanted: the image by f does not depend on the choice of g , and if we choose two different g 's, their image falls into the same coset.

The surjectivity is immediate.

We conclude the proof by showing the injectivity. Let us assume that $f(y_1) = f(y_2)$ for $y_1 = g_1 \cdot x \in B(x)$, $y_2 = g_2 \cdot x \in B(x)$. Thus

$$g_1\text{Stab}(x) = g_2\text{Stab}(x) \iff g_2^{-1}g_1 \in \text{Stab}(x) \iff g_2^{-1}g_1 \cdot x = x \iff g_1 \cdot x = g_2 \cdot x.$$

□

Let G be a finite group. We consider again as action the conjugation ($X = G$), given by: $g \cdot x = gxg^{-1}$. Recall that orbits under this action are given by

$$B(x) = \{gxg^{-1}, g \in G\}.$$

Let us notice that x always is in its orbit $B(x)$ (take $g = 1$). Thus if we have an orbit of size 1, this means that

$$gxg^{-1} = x \iff gx = xg$$

and we get an element x in the center $Z(G)$ of G . In words, elements that have an orbit of size 1 under the action by conjugation are elements of the center.

Recall that the orbits $B(x)$ partition X :

$$X = \sqcup B(x)$$

where the disjoint union is over a set of representatives. We get

$$\begin{aligned} |G| &= \sum |B(x)| \\ &= |Z(G)| + \sum |B(x)| \\ &= |Z(G)| + \sum [G : \text{Stab}(x)], \end{aligned}$$

where the second equality comes by splitting the sum between orbits with 1 element and orbits with at least 2 elements, while the third follows from the Orbit-Stabilizer Theorem. By remembering that $\text{Stab}(x) = C_G(x)$ when the action is the conjugation, we can alternatively write

$$|G| = |Z(G)| + \sum [G : C_G(x)].$$

This formula is called the [class equation](#).

Example 1.25. Consider the dihedral D_4 of order 8, given by

$$D_4 = \{1, s, r, r^2, r^3, rs, r^2s, r^3s\},$$

with $s^2 = 1$, $r^4 = 1$ and $sr s = r^{-1}$. We have that the center $Z(D_4)$ of D_4 is $\{1, r^2\}$ (just check that $r^2s = sr^2$). There are three conjugacy classes given by

$$\{r, r^3\}, \{rs, r^3s\}, \{s, r^2s\}.$$

Thus

$$|D_4| = 8 = |Z(D_4)| + |B(r)| + |B(rs)| + |B(s)|.$$

The following result has many names: Burnside's lemma, Burnside's counting theorem, the Cauchy-Frobenius lemma or the orbit-counting theorem. This result is not due to Burnside himself, who only quoted it. It is attributed to Frobenius.

Theorem 1.25. (Orbit-Counting Theorem). *Let the finite group G act on the finite set X , and denote by X^g the set of elements of X that are fixed by g , that is $X^g = \{x \in X, g \cdot x = x\}$. Then*

$$\text{number of orbits} = \frac{1}{|G|} \sum_{g \in G} |X^g|,$$

that is the number of orbits is the average number of points left fixed by elements of G .

Proof. We have

$$\begin{aligned} \sum_{g \in G} |X^g| &= |\{(g, x) \in G \times X, g \cdot x = x\}| \\ &= \sum_{x \in X} |\text{Stab}(x)| \\ &= \sum_{x \in X} |G|/|B(x)| \end{aligned}$$

by the Orbit-Stabilizer Theorem. We go on:

$$\begin{aligned} \sum_{x \in X} |G|/|B(x)| &= |G| \sum_{x \in X} 1/|B(x)| \\ &= |G| \sum_{B \in \text{set of orbits}} \sum_{x \in B} \frac{1}{|B|} \\ &= |G| \sum_{B \in \text{set of orbits}} 1 \end{aligned}$$

which concludes the proof. Note that the second equality comes from the fact that we can write X as a disjoint union of orbits. \square

1.8 The Sylow theorems

We look at orders of groups again, but this time paying attention to the occurrence of prime factors. More precisely, we will fix a given prime p , look at the partial factorization of the group order n as $n = p^r m$ where p does not divide m , and study the existence of subgroups of order p or a power of p . In a sense, this is trying to establish some kind of converse for Lagrange's Theorem. Recall that Lagrange's Theorem tells that the order of a subgroup divides the order of the group. Here we conversely pick a divisor of the order of the group, and we try to find a subgroup with order the chosen divisor.

Definition 1.25. Let p be a prime. The group G is said to be a p -group if the order of each element of G is a power of p .

Examples 1.26. We have already encountered several 2-groups.

1. We have seen in Example 1.14 that the cyclic group C_4 has elements of order 1, 2 and 4, while the direct product $C_2 \times C_2$ has elements of order 1 and 2.
2. The dihedral group D_4 is also a 2-group.

Definition 1.26. If $|G| = p^r m$, where p does not divide m , then a subgroup P of order p^r is called a **Sylow p -subgroup** of G . Thus P is a p -subgroup of G of maximum possible size.

The first thing we need to check is that such a subgroup of order p^r indeed exists, which is not obvious. This will be the content of the first Sylow theorem. Once we have proven the existence of a subgroup of order p^r , it has to be a p -group, since by Lagrange's Theorem the order of each element must divide p^r .

We need a preliminary lemma.

Lemma 1.26. If $n = p^r m$ where p is prime, then $\binom{n}{p^r} \equiv m \pmod{p}$. Thus if p does not divide m , then p does not divide $\binom{n}{p^r}$.



Figure 1.3: Ludwig Sylow (1832-1918)

Proof. We have to prove that

$$\binom{n}{p^r} \equiv m \pmod{p},$$

after which we have that if p does not divide m , the $m \not\equiv 0 \pmod{p}$ implying that $\binom{n}{p^r} \not\equiv 0 \pmod{p}$ and thus p does not divide $\binom{n}{p^r}$.

Let us use the binomial expansion of the following polynomial

$$(x+1)^{p^r} = \sum_{k=0}^{p^r} \binom{p^r}{k} x^{p^r-k} 1^k \equiv x^{p^r} + 1 \pmod{p}$$

where we noted that all binomial coefficients but the first and the last are divisible by p . Thus

$$(x+1)^{p^r m} \equiv (x^{p^r} + 1)^m \pmod{p}$$

which we can expand again into

$$\sum_{k=0}^{p^r m} \binom{p^r m}{k} x^{p^r m - k} \equiv \sum_{k=0}^m \binom{m}{k} (x^{p^r})^{m-k} \pmod{p}.$$

We now look at the coefficient of x^{p^r} on both sides:

- on the left, take $k = p^r(m-1)$, to get $\binom{p^r m}{p^r(m-1)}$,
- on the right, take $k = m-1$, to get $\binom{m}{m-1} = m$.

The result follows by identifying the coefficients of x^{p^r} . □

We are ready to prove the first Sylow Theorem.

Theorem 1.27. (1st Sylow Theorem). *Let G be a finite group of order $p^r m$, p a prime such that p does not divide m , and r some positive integer. Then G has at least one Sylow p -subgroup.*

Proof. The idea of the proof is to actually exhibit a subgroup of G of order p^r . For that, we need to define a clever action of G on a carefully chosen set X . Take the set

$$X = \{\text{subsets of } G \text{ of size } p^r\}$$

and for action that G acts on X by left multiplication. This is clearly a well-defined action. We have that

$$|X| = \binom{p^r m}{p^r}$$

which is not divisible by p (by the previous lemma). Recall that the action of G on X induces a partition of X into orbits:

$$X = \sqcup B(S)$$

where the disjoint union is taken over a set of representatives. Be careful that here S is an element of X , that is S is a subset of size p^r . We get

$$|X| = \sum |B(S)|$$

and since p does not divide $|X|$, it does not divide $\sum |B(S)|$, meaning that there is at least one S for which p does not divide $|B(S)|$. Let us pick this S , and denote by P its stabilizer.

The subgroup P which is thus by choice the stabilizer of the subset $S \in X$ of size p^r whose orbit size is not divisible by p is our candidate: we will prove it has order p^r .

$|P| \geq p^r$. Let us use the Orbit-Stabilizer Theorem, which tells us that

$$|B(S)| = |G|/|P| = p^r m/|P|.$$

By choice of the S we picked, p does not divide $|B(S)|$, that is p does not divide $p^r m/|P|$ and $|P|$ has to be a multiple of p^r , or equivalently p^r divides $|P|$.

$|P| \leq p^r$. Let us define the map λ_x , $x \in S$, by

$$\lambda_x : P \rightarrow S, \quad g \mapsto \lambda_x(g) = gx.$$

In words, this map goes from P , which is a subgroup of G , to S , which is an element of X , that is a subset of G with cardinality p^r . Note that this map is well-defined since $gx \in S$ for any $x \in S$ and any $g \in P$ by definition of P being the stabilizer of S . It is also clearly injective ($gx = hx$ implies $g = h$ since x is an element of the group G and thus is invertible). If we have an injection from P to S , that means $|P| \leq |S| = p^r$.



Figure 1.4: Augustin Louis Cauchy (1789-1857)

□

Corollary 1.28. (Cauchy Theorem). *If the prime p divides the order of G , then G has an element of order p .*

Proof. Let P be a Sylow p -subgroup of G (which exists by the 1st Sylow Theorem), and pick $x \neq 1$ in P . The order $|x|$ of x is a power of p by definition of a p -group, say $|x| = p^k$. Then $x^{p^{k-1}}$ has order p . □

The above corollary gives some converse to Lagrange's Theorem. The one below gives an alternative definition of a finite p -group. It is tempting to use it as a definition of p -group, however it cannot be used for infinite groups.

Corollary 1.29. *A finite group G is a p -group if and only if the order of G is a power of p .*

Proof. If the order of G is not a power of p , then it is divisible by some other prime q , in which case G contains an element of order q by Cauchy's Theorem, which contradicts the definition of p -group.

The converse is clear using Lagrange's Theorem. □

Now that we know that at least one Sylow p -subgroup exists, let us derive a result on the number n_p of Sylow p -subgroups in a group G .

We need again a preliminary lemma.

Lemma 1.30. *Let H and K be arbitrary finite subgroups of a group G . Then*

$$|HK| = \frac{|H||K|}{|H \cap K|}.$$

Note that if K is further assumed to be normal in G , then this result can be deduced from the second isomorphism Theorem.

Proof. Consider the map

$$f : H \times K \rightarrow HK, (h, k) \mapsto hk.$$

Since f is surjective, $|HK| \leq |H \times K| < \infty$ since H and K are finite, and thus HK is finite. Let h_1k_1, \dots, h_dk_d be the distinct elements of HK . Then $H \times K$ is the disjoint union of the $f^{-1}(h_ik_i)$, $i = 1, \dots, d$. Now we can check that

$$f^{-1}(hk) = \{(hg, g^{-1}k), g \in H \cap K\}$$

and this set has cardinality

$$|f^{-1}(hk)| = |H \cap K|.$$

Thus

$$|H \times K| = d|H \cap K|$$

which concludes the proof. \square

Theorem 1.31. (2nd Sylow Theorem). *Let G be a finite group of order $p^r m$, p a prime such that p does not divide m , and r some positive integer. Denote by n_p the number of Sylow p -subgroups of G . Then*

$$n_p \equiv 1 \pmod{p}.$$

Proof. Consider the set

$$X = \{ \text{all Sylow } p\text{-subgroups of } G \}$$

whose cardinality $|X|$ is denoted by n_p . By the 1st Sylow Theorem, this set is non-empty and there exists at least one Sylow p -subgroup P in X , whose order is p^r . We can thus let P act on X by conjugation, i.e., $g \cdot Q = gQg^{-1}$, $g \in P$, $Q \in X$. Note that in the case where P is the only Sylow p -subgroup, then we can take $Q = P$.

By the Orbit-Stabilizer Theorem, we have that the orbit $B(Q)$ of any Sylow p -subgroup Q in X has cardinality

$$|B(Q)| = |P|/|\text{Stab}(Q)| = p^r/|\text{Stab}(Q)|.$$

In particular, the size of an orbit of any Sylow p -subgroups divides p^r , meaning it has to be either 1 or a power of p .

Let us recall that the set X is partitioned by the orbits $B(Q)$ under the action of P , so that the cardinality of X is:

$$|X| = \sum |B(Q)| = \sum |B(Q')| + \sum |B(Q'')|$$

where Q' and Q'' denote subgroups whose orbit has respectively one element or at least two elements. Since p divides the second sum, we have

$$|X| \equiv \text{number of orbits of size 1} \pmod{p}.$$

To conclude the proof, we thus have to show that there is only one Sylow p -subgroup whose orbit has size 1, namely P itself (it is clear that P has an orbit of size 1, since conjugating P by itself will not give another subgroup).

Let us assume there is another Sylow p -subgroup Q whose orbit has only one element, namely (recall that one element is always in its orbit):

$$gQg^{-1} = Q, \quad g \in P,$$

which translates into

$$gQ = Qg \text{ for all } g \in P \iff PQ = QP.$$

This easily implies that PQ is a subgroup of G , and by the previous lemma

$$|PQ| = \frac{p^r p^r}{|P \cap Q|}$$

implying that $|PQ|$ is a power of p , say p^c for some c which cannot be bigger than r , since $|G| = p^r m$. Thus

$$p^r = |P| \leq |PQ| \leq p^r$$

so that $|PQ| = p^r$ and thus $|P| = |PQ|$, saying that Q is included in P . But both Q and P have same cardinality of p^r , so $Q = P$. \square

The third of the Sylow Theorems tells us that all Sylow p -subgroups are conjugate.

Theorem 1.32. (3rd Sylow Theorem). *Let G be a finite group of order $p^r m$, p a prime such that p does not divide m , and r some positive integer. Then all Sylow p -subgroups are conjugate.*

Proof. Let P be a Sylow p -subgroup of G and let R be a p -group of G . We will prove that R (being a p -group in general) is contained in a conjugate of P .

Let R act by multiplication on the set Y of left cosets of P :

$$Y = \{gP, g \in G\}.$$

It is a well-defined action (it is multiplication in the group G).

We want to prove that there is an orbit of size 1 under this action. By Lagrange's Theorem, we know that

$$|Y| = |G|/|P| = \frac{p^r m}{p^r} = m$$

and thus p does not divide $|Y|$ by assumption on m . By writing that we have a partition of Y by its orbits, we get

$$|Y| = \sum |B(y)|$$

and there exists one orbit $B(y)$ whose size is not divisible by p . By the Orbit-Stabilizer Theorem, we have that the size of every orbit divides $|R|$, which has order a power of p (by a corollary of the 1st Sylow Theorem), so every orbit size must divide p , which gives as only possibility that there is an orbit of size 1.

Let $gP \in Y$ be the element whose orbit size is 1. We have

$$h \cdot gP = gP$$

for $h \in R$, since gP belongs to its orbit. Thus

$$g^{-1}hg \in P \iff h \in gPg^{-1}$$

for all $h \in R$. We have just proved that the p -group R is contained in a conjugate of P .

All we needed for the proof is that R is a p -group, so the same proof holds for the case of a Sylow p -subgroup, for which we get that R is contained in a conjugate of P , and both have same cardinality, which concludes the proof.

We will use the fact that the proof works for R a p -group in general for proving one corollary. \square

Corollary 1.33. 1. Every p -subgroup of G is contained in a Sylow p -subgroup.
2. The number n_p of Sylow p -subgroups divides m .

Proof. 1. Now we know that if P is a Sylow p -subgroup, then so is gPg^{-1} , $g \in G$, by the above theorem. The proof of the theorem itself shows that any p -group is included in gPg^{-1} and we are done.

2. Let the group G act by conjugation on the set of its subgroups. In particular, G acts on the Sylow p -subgroup P , and the orbit of P has size the number of Sylow p -subgroups in G , denoted by n_p . By the Orbit-Stabilizer Theorem, n_p divides $|G| = p^r m$. But p cannot be a prime factor of n_p since $n_p \equiv 1 \pmod{p}$, from which it follows that n_p must divide m . \square

1.9 Simple groups

We will now see a few applications of the Sylow Theorems, in particular to determine the structure of so-called simple groups.

Definition 1.27. A group G is **simple** if $G \neq \{1\}$ and the only normal subgroups of G are G itself and $\{1\}$.

Finite simple groups are important because in a sense they are building blocks of all finite groups, similarly to the way prime numbers are building blocks of the integers. This will be made clearer in the coming section by the Jordan-Hölder Theorem. Infinite groups exist and can be found among Lie groups, but we will concentrate here on finite groups.

The case of simple abelian groups is easy to understand. Suppose that G is a simple abelian group. Note that G cannot be $\{1\}$. Now G being abelian, all its subgroups are normal. On the other hand, G being simple, its only normal subgroups are $\{1\}$ and itself, leaving as only solution that G has only two subgroups, namely $\{1\}$ and G . Thus G has to be a cyclic group of prime order.

We now start looking at non-abelian simple groups. We start with some preliminary results.

Proposition 1.34. *If P is a non-trivial finite p -group, then P has a non-trivial center.*

Proof. Let P act on itself by conjugation. The orbits of this action are the conjugacy classes of P , and we have that x belongs to an orbit of size 1 if and only if x belongs to the center $Z(P)$.

By the Orbit-Stabilizer, the size of any orbit must divide $|P|$, which is a power of p by a corollary of the 1st Sylow Theorem.

If it were true that the center is trivial, that is $Z(P) = \{1\}$, then that means there is only one orbit of size 1, and thus all the other orbits must have size that divides p , namely they are congruent to 0 mod p . Thus

$$|P| = |Z(P)| + \sum |B| \equiv 1 \pmod{p},$$

where the sum is over orbits of size at least 2. This is clearly a contradiction, which concludes the proof. \square

Lemma 1.35. *The group P is a normal Sylow p -subgroup if and only if P is the unique Sylow p -subgroup of G .*

Proof. We know from the 3rd Sylow Theorem that the Sylow p -subgroups form a single conjugacy class. Then P is the unique Sylow p -subgroup means that P is the only element in the conjugacy class, and thus it satisfies

$$gPg^{-1} = P$$

for every $g \in G$, which exactly means that P is a normal subgroup of G . \square

Thanks to the two above results, we can now prove that a non-abelian simple group must have more than one Sylow p -subgroup.

Proposition 1.36. *Let G be a finite group which is non-abelian and simple. If the prime p divides $|G|$, then the number n_p of Sylow p -subgroups is strictly bigger than 1.*

Proof. Let us look at the prime factors appearing in the order of G .

- If p is the only prime factor of $|G|$, then $|G|$ must be a power of p , that is G is a non-trivial p -group (it is non-trivial by definition of simple and a p -group by a corollary of the 1st Sylow Theorem). Now the above proposition tells us that its center $Z(G)$ is non-trivial as well. Since $Z(G)$ is a normal subgroup of G and G is simple, it must be that $G = Z(G)$, which contradicts the assumption that G is non-abelian.

- We then know that $|G|$ is divisible by at least two distinct primes. So if P is a Sylow p -subgroup, then

$$\{1\} < P < G,$$

where the second inclusion is strict since the order of G is divisible by two primes.

If there were only one Sylow p -subgroup, namely $n_p = 1$, then this Sylow p -subgroup would be normal by the above lemma, which contradicts the simplicity of G .

□

Let us see if we can be more precise by refining the assumptions on the order of the group G we consider. The group G can be either abelian or non-abelian, though the results on simplicity are more interesting for non-abelian groups.

Proposition 1.37. *Let G be a group of order pq , where p and q are distinct primes.*

1. *If $q \not\equiv 1 \pmod{p}$, then G has a normal Sylow p -subgroup.*
2. *If both $q \not\equiv 1 \pmod{p}$ and $p \not\equiv 1 \pmod{q}$, then G is cyclic.*
3. *G is not simple.*

Proof. 1. By Lemma 1.35, saying that G has a normal Sylow p -subgroup is the same as saying that there is a unique Sylow p -subgroup. This is now indeed the case, since the number n_p of Sylow p -subgroups has to satisfy both $n_p \equiv 1 \pmod{p}$ and $n_p \mid q$ by the Sylow Theorems. Since q is prime, n_p is either 1 or q . It cannot be that $n_p = q$, since it would imply that $q \equiv 1 \pmod{p}$ which contradicts the assumption.

2. By the previous point, the group G has a normal Sylow p -subgroup P and a normal Sylow q -subgroup Q , both of them cyclic (since they are of prime order). Let us write them respectively $P = \langle x \rangle$ and $Q = \langle y \rangle$. Since both P and Q are normal, with $P \cap Q = \{1\}$, we have that $xy = yx$ (we have seen that before, but the argument goes like that: take the element $xyx^{-1}y^{-1}$ and show that it belongs to $P \cap Q$ by normality of P and Q). Thanks to this commutativity property, we have that $(xy)^n = x^n y^n$ and the order of xy is pq , showing that G is cyclic with generator xy .
3. Without loss of generality, we can assume that $p > q$ so that p does not divide $q - 1$ which can be rewritten as

$$q \not\equiv 1 \pmod{p}.$$

By the first point, we know that G has a normal Sylow p -group, and thus G cannot be simple.

□

Here is another family of groups which are not simple. The proof contains an interesting combinatorial argument!

Proposition 1.38. *Let G be a group of order $|G| = p^2q$ where p and q are two distinct primes. Then G contains either a normal Sylow p -subgroup or a normal Sylow q -subgroup. In particular, G is not simple.*

Proof. Recall that having a normal Sylow p -subgroup (resp. q -subgroup) is the same as saying there is a unique Sylow p -subgroup (resp. q -subgroup). Suppose that the claim is not true, that is both the number of Sylow p -subgroups n_p and the number of Sylow q -subgroups n_q are bigger than 1. Let us start this proof by counting the number of elements of order q in G .

If a Sylow q -subgroup has order q , it is cyclic and can be generated by any of its elements which is not 1. This gives $q - 1$ elements of order q per Sylow q -subgroup of G . Conversely, if y has order q , then the cyclic group it generates is a Sylow q -subgroup, and any two distinct Sylow q -subgroups have trivial intersection. Thus

$$\text{number of elements of order } q = n_q(q - 1).$$

Now we know from the Sylow Theorems that $n_q \mid p^2$, thus n_q is either p or p^2 ($n_q = 1$ is ruled out by the fact that we do a proof by contradiction).

- $n_q = p^2$: then the number of elements of order NOT q is

$$p^2q - p^2(q - 1) = p^2.$$

On the other hand, if P is a Sylow p -subgroup, then it also contains p^2 elements, and all of them have order not q , so that we can conclude that P actually contains all elements of order not q , which implies that we have only one Sylow p -subgroup, yielding the wanted contradiction.

- $n_q = p$: We know from Sylow Theorems that

$$n_q \equiv 1 \pmod{q} \Rightarrow p \equiv 1 \pmod{q} \Rightarrow p > q,$$

but also that

$$n_p \mid q$$

and since q is prime, that leaves $n_p = 1$ or $n_p = q$ and thus $n_p = q$. As before

$$n_p \equiv 1 \pmod{p} \Rightarrow q \equiv 1 \pmod{p} \Rightarrow q > p.$$

This concludes the proof. □

We have thus shown that the situation is easy for simple abelian groups. For non-abelian groups, we have seen two cases ($|G| = pq$ and $|G| = p^2q$) where groups are not simple. To find a non-abelian group which is simple, one has to

	simple	not simple
abelian	C_p	other groups
non-abelian	$n_p > 1$	$ G = pq, G = p^2q$

Table 1.3: C_p refers to a cyclic group of prime order, and n_p counts the number of Sylow p -subgroups. The groups of order pq and p^2q are also not simple for abelian groups and are included in other groups.

go to groups of order at least 60. Indeed, it has been proven that the smallest non-abelian simple group is the alternating group A_5 of order 60, this is the group of even permutations of a finite set. This result is attributed to Galois (1831). It is not an easy task to determine the list of simple groups, and in fact, the classification of finite simple groups was only accomplished in 1982 (there has been some controversy as to whether the proof is correct, given its length - tens of thousands of pages - and complexity).

1.10 The Jordan-Hölder Theorem

We have mentioned when introducing simple groups in the previous section that they can be seen as building blocks for decomposing arbitrary groups. This will be made precise in this section.

Definition 1.28. Let G be a group, and let G_0, \dots, G_n be subgroups of G such that

1. $G_n = \{1\}$ and $G_0 = G$,
2. $G_{i+1} \trianglelefteq G_i, i = 0, \dots, n-1$.

Then the series

$$\{1\} = G_n \trianglelefteq G_{n-1} \trianglelefteq \dots \trianglelefteq G_0 = G$$

is called a [subnormal series](#) for G .

Suppose that G_{i+1} is not a maximal normal subgroup of G_i , then we can refine the subnormal series by inserting a group H such that $G_{i+1} \triangleleft H \triangleleft G_i$, and we can repeat this process hoping it will terminate (it will if G is finite, it may not otherwise).

Definition 1.29. Let G be a group, and let G_0, \dots, G_n be subgroups of G such that

1. $G_n = \{1\}$ and $G_0 = G$,
2. $G_{i+1} \triangleleft G_i, i = 0, \dots, n-1$, such that G_{i+1} is a maximal normal subgroup of G_i .

Then the series

$$\{1\} = G_n \triangleleft G_{n-1} \triangleleft \cdots \triangleleft G_0 = G$$

is called a **composition series** for G . The factor groups G_i/G_{i+1} are called the **factors** of the composition series, whose length is n .

Another way of stating the condition G_{i+1} is a maximal normal subgroup of G_i is to say that G_i/G_{i+1} is simple, $i = 0, \dots, n-1$. To see that asks a little bit of work. This result is sometimes called the 4rth isomorphism theorem.

Theorem 1.39. (Correspondence Theorem). *Let N be a normal subgroup of G and let H be a subgroup of G containing N . Then the map*

$$\psi : \{\text{subgroups of } G \text{ containing } N\} \rightarrow \{\text{subgroups of } G/N\}, \quad H \mapsto \psi(H) = H/N$$

is a bijection. Furthermore, H is a normal subgroup of G if and only if H/N is a normal subgroup of G/N .

Proof. We first prove that ψ is a bijection.

Injectivity. If $H_1/N = H_2/N$, then cosets in each subgroup are the same, that is for any $h_1 \in H_1$, we have $h_1N = h_2N$ for some $h_2 \in H_2$, implying that $h_2^{-1}h_1 \in N \subset H_2$ and thus $h_1 \in H_2$, showing that $H_1 \subseteq H_2$. By repeating the same argument but reverting the role of H_1 and H_2 , we get $H_2 \subseteq H_1$ and thus $H_1 = H_2$.

Surjectivity. Let Q be a subgroup of G/N and let $\pi : G \rightarrow G/N$ be the canonical projection. Then

$$\pi^{-1}(Q) = \{a \in G, aN \in Q\}.$$

This is a subgroup of G containing N and

$$\psi(\pi^{-1}(Q)) = \{aN, aN \in Q\} = Q.$$

We are left to prove that $H \trianglelefteq G \iff H/N \trianglelefteq G/N$. Assume thus that $H \trianglelefteq G$. For any $a \in G$, we have to show that

$$(aN)(H/N)(aN)^{-1} = H/N.$$

Now for any $hN \in H/N$, we have

$$(aN)(hN)(aN)^{-1} = (aha^{-1})N \in H/N$$

and we are done.

Conversely, suppose that $H/N \trianglelefteq G/N$. Consider the homomorphism

$$a \mapsto (aN)(H/N)$$

which is the composition of the canonical projection π of G onto G/N , and the canonical projection of G/N onto $(G/N)/(H/N)$ (the latter makes sense since

$H/N \trianglelefteq G/N$). We now want to show that H is the kernel of this map, which will conclude the proof since the kernel of a group homomorphism is normal.

An element a is in the kernel if and only if $(aN)(H/N) = H/N$, that is if and only if $aN \in H/N$, or equivalently $aN = hN$ for some $h \in H$. Since N is contained in H , this means aN is in H and thus so is a , which is what we wanted to prove. \square

Let us now go back to the composition series of G . If G/N is simple, then by definition it has only trivial normal subgroups, namely N and G/N . Now using the Correspondence Theorem, the normal subgroups N and G/N exactly correspond to the normal subgroups N and G in G , which shows that N is the maximal normal subgroup of G .

The Jordan-Hölder Theorem will tell us that if G has a composition series, then the resulting composition length n and the simple composition factors G_i/G_{i+1} are unique up to isomorphism and rearrangement. This for example shows that if G_1 and G_2 are two groups with different composition factors, then they cannot be isomorphic.

Lemma 1.40. *Let G be a group with composition series*

$$\{1\} = G_n \triangleleft G_{n-1} \triangleleft \cdots \triangleleft G_0 = G.$$

Then for any normal subgroup K of G , if we remove the duplicates from the series

$$\{1\} = K \cap G_n \trianglelefteq K \cap G_{n-1} \trianglelefteq \cdots \trianglelefteq K \cap G_0 = K,$$

the result is a composition series for K of length at most n .

Proof. We need to show that $K \cap G_{i+1} \triangleleft K \cap G_i$ and that the group $(K \cap G_i)/(K \cap G_{i+1})$ is simple for all i .

Let $x \in K \cap G_i$ and $g \in K \cap G_{i+1}$. Then $xgx^{-1} \in K$ since by assumption K is a normal subgroup of G , and $xgx^{-1} \in G_{i+1}$ since $G_{i+1} \triangleleft G_i$. Thus $xgx^{-1} \in K \cap G_{i+1}$ which proves that $K \cap G_{i+1} \triangleleft K \cap G_i$.

We now look at the quotient group $(K \cap G_i)/(K \cap G_{i+1})$. Since G_i/G_{i+1} is simple, G_{i+1} is a maximal normal subgroup of G_i , and thus the only normal subgroups of G_i that contain G_{i+1} are G_i and G_{i+1} .

Recall that $K \cap G_i$ is normal in G_i (it is the kernel of the canonical projection of G to G/K restricted to G_i), so that we get

$$G_{i+1} \triangleleft (K \cap G_i)G_{i+1} \triangleleft G_i.$$

For the first normal inclusion, compute that for $kg \in (K \cap G_i)G_{i+1}$ we have

$$kgG_{i+1}g^{-1}k^{-1} = kG_{i+1}k^{-1} \subseteq G_{i+1}$$

since $k \in G_i$ and G_{i+1} is normal in G_i . For the second normal inclusion, we have for $g \in G_i$ that

$$g(K \cap G_i)G_{i+1}g^{-1} = (K \cap G_i)gG_{i+1}g^{-1}$$

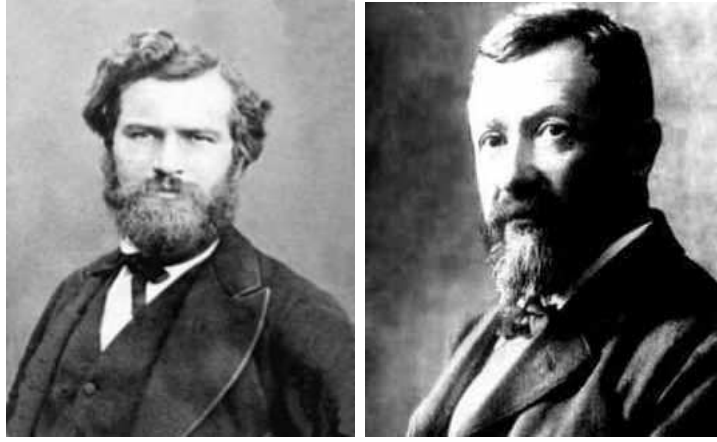


Figure 1.5: Camille Jordan (1838–1922) and Otto Hölder (1859–1937)

since $K \cap G_i$ is normal in G_i and

$$(K \cap G_i)gG_{i+1}g^{-1} \subseteq (K \cap G_i)G_{i+1}$$

since $G_{i+1} \triangleleft G_i$.

Thus either $G_{i+1} = (K \cap G_i)G_{i+1}$ or $(K \cap G_i)G_{i+1} = G_i$. Using the second isomorphism theorem (with $G_{i+1} \triangleleft G_i$ and $(K \cap G_i) \leq G_i$), we have

$$(K \cap G_i)G_{i+1}/G_{i+1} \simeq (K \cap G_i)/(K \cap G_i \cap G_{i+1}) = (K \cap G_i)/(K \cap G_{i+1}).$$

We can see that if $G_{i+1} = (K \cap G_i)G_{i+1}$, then $K \cap G_i = K \cap G_{i+1}$ and we have a duplicate to remove. If $(K \cap G_i)G_{i+1} = G_i$, then

$$G_i/G_{i+1} \simeq (K \cap G_i)/(K \cap G_{i+1})$$

and thus $(K \cap G_i)/(K \cap G_{i+1})$ is simple. \square

Theorem 1.41. (Jordan-Hölder Theorem). *Let G be a group that has a composition series. Then any two composition series for G have the same length. Moreover, if*

$$\{1\} = G_n \triangleleft G_{n-1} \triangleleft \cdots \triangleleft G_0 = G$$

and

$$\{1\} = H_n \triangleleft H_{n-1} \triangleleft \cdots \triangleleft H_0 = G$$

are two composition series for G , there exists a permutation τ such that $G_i/G_{i+1} \simeq H_{\tau(i)}/H_{\tau(i)+1}$.

Proof. The proof will be on induction on the length of a composition series. Suppose that G is a group with a composition series of length 1. Then the subnormal series

$$G \triangleright \{1\}$$

cannot be refined, so it must be a composition series. In particular $G \simeq G/\{1\}$ is simple. This is also the only composition series for G and so all the assertions are true for length 1.

Suppose now that $n > 1$ and that the claims are true for composition series of length up till $n - 1$. Let G be a group with composition series of length n , say

$$\{1\} = G_n \triangleleft G_{n-1} \triangleleft \cdots \triangleleft G_0 = G$$

(so that $G_i \neq G_{i+1}$ for each i). Now let

$$\{1\} = H_m \triangleleft H_{m-1} \triangleleft \cdots \triangleleft H_0 = G$$

be a composition series for G (again $H_i \neq H_{i+1}$ for each i).

We first have to show that $m = n$ after which we discuss the unicity of the decomposition.

(Proof that $m = n$). The idea of the proof goes as follows: to use the induction hypothesis, we need to get a composition series of length smaller than n , that is, we need to identify the first composition factors, which we will use the above lemma. Concretely, we first exclude the case when $G_1 = H_1$, then compute a composition series of length $n - 2$ for $H_1 \cap G_1$, which will indeed be the second composition factor. We then use the second composition series of G to get another composition series for $H_1 \cap G_1$ whose length depends on m , that we can compare to the known one.

If $G_1 = H_1$, then by the induction hypothesis applied to G_1 , we have $n - 1 = m - 1$, we have a suitable permutation τ of the $n - 1$ factors, and we are done.

Suppose then that $H_1 \neq G_1$. Since both G_1 and H_1 are maximal normal in G , we see that $H_1 \triangleleft G_1 H_1 \triangleleft G$ with $H_1 \neq G_1 H_1$ since we assumed $H_1 \neq G_1$. Thus $G_1 H_1 = G$, from which we conclude by the 2nd isomorphism theorem that

$$G_1 H_1 / H_1 \simeq G / H_1 \simeq G_1 / (H_1 \cap G_1).$$

Since G/H_1 is simple, we get that $G_1/(H_1 \cap G_1)$ is simple as well. Now by the above lemma, upon removing duplicates from the series

$$\{1\} = H_1 \cap G_n \trianglelefteq \cdots \trianglelefteq H_1 \cap G_0 = H_1,$$

we get a composition series for H_1 of length at most n and thus upon removing duplicates

$$\{1\} = H_1 \cap G_n \trianglelefteq \cdots \trianglelefteq H_1 \cap G_1$$

is a composition series for $H_1 \cap G_1$ of length at most $n - 1$. Since $G_1/(H_1 \cap G_1)$ is simple, it follows that upon removing duplicates

$$\{1\} = H_1 \cap G_n \trianglelefteq \cdots \trianglelefteq H_1 \cap G_1 \triangleleft G_1$$

is a composition series for G_1 . But then

$$G_1 \triangleright G_2 \triangleright \cdots \triangleright G_n = \{1\}$$

and

$$G_1 \triangleright H_1 \cap G_1 \supseteq H_1 \cap G_2 \supseteq \cdots \supseteq H_1 \cap G_n = \{1\}$$

are both composition series for G_1 , with the first series of length $n - 1$. By induction hypothesis, both series have the same length. Since $G_1 \neq H_1 \cap G_1$ (recall that we assumed $H_1 \neq G_1$), any duplication must occur later in the series. Let

$$G_1 = K_1 \triangleright K_2 = H_1 \cap G_1 \triangleright K_3 \triangleright \cdots \triangleright K_n = \{1\}$$

denote the composition series for G_1 of length $n - 1$ that results from removing the duplicates. By hypothesis, there exists a permutation α such that $G_i/G_{i+1} \simeq K_{\alpha(i)}/K_{\alpha(i)+1}$ for each $i = 1, \dots, n - 1$. Set α not to move the index 0, then

$$G = G_0 \triangleright G_1 \triangleright G_2 \triangleright \cdots \triangleright G_n = \{1\}$$

and

$$G = K_0 \triangleright G_1 = K_1 \triangleright K_2 = H_1 \cap G_1 \triangleright K_3 \triangleright \cdots \triangleright K_n = \{1\}$$

are composition series of length n for G and α is a permutation such that $G_i/G_{i+1} \simeq K_{\alpha(i)}/K_{\alpha(i)+1}$ for each $i = 0, \dots, n - 1$. Moreover, we have found a composition series for $H_1 \cap G_1$ of length $n - 2$.

Let us now repeat similar computations for the composition series

$$G = H_0 \triangleright H_1 \triangleright \cdots \triangleright H_m = \{1\}$$

and the normal subgroup G_1 of G . Again by the above lemma, upon removing the duplicates from the series

$$G_1 = H_0 \cap G_1 \supseteq H_1 \cap G_1 \supseteq \cdots \supseteq H_m \cap G_1 = \{1\}$$

we obtain a composition series for G_1 , so that upon removing the duplicates

$$H_1 \cap G_1 \supseteq \cdots \supseteq H_m \cap G_1 = \{1\}$$

yields a composition series for $H_1 \cap G_1$. Now since $H_1 \cap G_1$ has a composition series of length $n - 2$, namely

$$K_2 = H_1 \cap G_1 \triangleright \cdots \triangleright K_n = \{1\},$$

we apply the induction hypothesis to $H_1 \cap G_1$ to conclude that all composition series of $H_1 \cap G_1$ have length $n - 2$, and so in particular the preceding composition series

$$H_1 \cap G_1 \supseteq \cdots \supseteq H_m \cap G_1 = \{1\}$$

has length $n - 2$. We cannot conclude yet, since we do not know how many terms there are in function of m in the above composition series (we need to get rid of the duplicates).

Since we know from the 2nd isomorphism theorem that $H_1/(H_1 \cap G_1) \simeq H_1G_1/G_1 = G_0/G_1$, which is a simple group, it follows that $H_1/(H_1 \cap G_1)$ is simple. Thus upon the removal of the duplicates from

$$H_1 \triangleright H_1 \cap G_1 \triangleright \cdots \triangleright H_m \cap G_1 = \{1\}$$

the result is a composition series for H_1 of length $n - 1$ (we added the term H_1 to the composition series for $H_1 \cap G_1$ of length $n - 2$). Also

$$H_1 \triangleright H_2 \triangleright \cdots \triangleright H_m = \{1\}$$

is another composition series for H_1 . Since the first series has length $n - 1$, by our induction hypothesis, the second series must also have length $n - 1$. Since its length is $m - 1$, it follows that $m = n$.

(Unicity of the composition factors). Again by induction hypothesis on H_1 , we have a permutation β of the $n - 1$ composition factors (which can be extended to n factors by setting $\beta(0) = 0$.) Namely, let L_i , $i = 1, 2, \dots, n$ denote the distinct terms in the series

$$H_1 \triangleright H_1 \cap G_1 \triangleright H_2 \cap G_1 \triangleright \cdots \triangleright H_n \cap G_1 = \{1\}$$

so that $L_1 = H_1$ and $L_2 = H_1 \cap G_1$. Then we have composition series

$$G = H_0 \triangleright H_1 \triangleright \cdots \triangleright H_n = \{1\} \text{ and } G = L_0 \triangleright L_1 \triangleright \cdots \triangleright L_n = \{1\}$$

of length n for G and there exists a permutation β of $\{0, 1, \dots, n - 1\}$ such that $H_i/H_{i+1} \simeq L_{\beta(i)}/L_{\beta(i)+1}$ for each $i = 0, 1, \dots, n - 1$.

We are almost done but for the fact that we need an isomorphism between H_i/H_{i+1} and $G_{\beta(i)}/G_{\beta(i)+1}$ instead of having $H_i/H_{i+1} \simeq L_{\beta(i)}/L_{\beta(i)+1}$. We recall that we already have a permutation α such that $G_i/G_{i+1} \simeq K_{\alpha(i)}/K_{\alpha(i)+1}$. We are thus left to find one between L_i/L_{i+1} and K_i/K_{i+1} .

Finally, since $K_2 = L_2 = H_1 \cap G_1$, we have two composition series for G :

$$\begin{array}{l} G \triangleright G_1 \triangleright H_1 \cap G_1 \triangleright K_3 \triangleright \cdots \triangleright K_{n-1} \triangleright K_n = \{1\} \\ G \triangleright H_1 \triangleright H_1 \cap G_1 \triangleright L_3 \triangleright \cdots \triangleright L_{n-1} \triangleright L_n = \{1\}. \end{array}$$

We may apply the induction hypothesis to $H_1 \cap G_1$ to obtain the existence of a permutation γ of $\{2, 3, \dots, n - 1\}$ such that for each i in this set we have $K_i/K_{i+1} \simeq L_{\gamma(i)}/L_{\gamma(i)+1}$. We have already seen that $G/G_1 \simeq H_1/(H_1 \cap G_1)$ and $G/H_1 \simeq G_1/(H_1 \cap G_1)$, so we may extend γ to a permutation of $\{0, 1, \dots, n - 1\}$ by setting $\gamma(0) = 1$ and $\gamma(1) = 0$. Then since

$$K_0 = G = L_0, \quad K_1 = G_1, \quad L_1 = H_1, \quad K_2 = L_2 = H_1 \cap G_1,$$

we have

$$K_i/K_{i+1} \simeq L_{\gamma(i)}/L_{\gamma(i)+1}, \quad i = 0, \dots, n - 1.$$

In summary, we have $m = n$, and for $\tau = \beta^{-1}\gamma\alpha$, we have

$$G_i/G_{i+1} \simeq H_{\tau(i)}/H_{\tau(i)+1}, \quad i = 0, \dots, n - 1.$$

This concludes the proof. \square

Example 1.27. The cyclic group C_{12} has three composition series

$$C_1 \triangleleft C_2 \triangleleft C_6 \triangleleft C_{12}, \quad C_1 \triangleleft C_2 \triangleleft C_4 \triangleleft C_{12}, \quad C_1 \triangleleft C_3 \triangleleft C_6 \triangleleft C_{12}$$

and all of them have the same length. Furthermore, the factor groups appearing are

$$\{C_2, C_3, C_2\}, \{C_2, C_2, C_3\}, \{C_3, C_2, C_2\}$$

which are indeed the same up to permutation.

Corollary 1.42. (Fundamental Theorem of arithmetic). *Let $n > 1$ be a positive integer. Then there exist unique primes $p_1 < p_2 < \dots < p_k$ and unique positive integers r_1, \dots, r_k such that $n = p_1^{r_1} p_2^{r_2} \dots p_k^{r_k}$.*

Proof. Let $G = \langle g \rangle$ be a cyclic group of order n . Then every subgroup of G is normal, and there is a unique subgroup of size d for each positive divisor d of n . Let d be the largest proper divisor of n , and let G_1 be the unique subgroup of G of size d . Then G/G_1 is simple and cyclic, hence of prime order. We may repeat this construction on the cyclic subgroup G_1 , so by induction, we obtain a composition series

$$G = G_0 \triangleleft G_1 \triangleleft G_2 \triangleleft \dots \triangleleft G_m = \{1\}$$

for G with G_i/G_{i+1} of prime order p_i for each i . Thus

$$\begin{aligned} n &= |G| \\ &= |G/G_1||G_1| \\ &= |G/G_1||G_1/G_2| \dots |G_{m-1}/G_m||G_m| \\ &= p_1 p_2 \dots p_{m-1}. \end{aligned}$$

The uniqueness of the prime decomposition of n follows from the Jordan-Hölder Theorem applied to G . \square

1.11 Solvable and nilpotent groups

Let us start by introducing a notion stronger than normality.

Definition 1.30. A subgroup H of the group G is called **characteristic** in G if for each automorphism f of G , we have

$$f(H) = H.$$

We may write $H \text{ char } G$.

This is stronger than normal since normality corresponds to choose for f the conjugation by an element of G .

Note that f restricted to H a characteristic subgroup (denoted by $f|_H$) is an automorphism of H (it is an endomorphism by definition of H being characteristic).

Here are a few immediate properties of characteristic subgroups.

Lemma 1.43. *Let G be a group, and let H, K be subgroups of G .*

1. *If H is characteristic in K and K is characteristic in G , then H is characteristic in G (being characteristic is transitive).*
2. *If H is characteristic in K , and K is normal in G , then H is normal in G .*

Proof. 1. Let ϕ be an automorphism of G . Since K is characteristic in G , then $\phi(K) = K$ by definition, and thus $\phi|_K$ is an automorphism of K . Now $\phi|_K(H) = H$ since H is characteristic in K . But $\phi|_K$ is just the restriction of ϕ , so $\phi(H) = H$.

2. Consider the automorphism of K given by $k \mapsto gkg^{-1}$, $g \in G$, which is well defined since K is normal in G . For any choice of g , we get a different automorphism of K , which will always preserve H since H is characteristic in K , and thus $gHg^{-1} \subset H$ which proves that H is normal in G . □

Let us introduce a new definition, that will give us an example of characteristic subgroup.

Definition 1.31. The **commutator subgroup** G' of a group G is the subgroup generated by all commutators

$$[x, y] = xyx^{-1}y^{-1}.$$

It is also called the **derived subgroup** of G .

Let us make a few remarks.

- By the subgroup generated by all commutators, we mean that by definition we take the smallest subgroup of G containing all the commutators. It is thus closed by construction.
- The product of two or more commutators need not be a commutator. It is known that the least order of a finite group for which there exists two commutators whose product is not a commutator is 96.
- Note that the inverse $[x, y]^{-1}$ of $[x, y]$ is given by $[x, y]^{-1} = [y, x] = yxy^{-1}x^{-1}$.

Here are a list of properties of the commutator subgroup G' .

Lemma 1.44. *Let G' be the commutator subgroup of G .*

1. *G' is characteristic in G .*
2. *G is abelian if and only if G' is trivial.*
3. *G/G' is abelian.*

4. If N is normal in G , then G/N is abelian if and only if $G' \leq N$.

Proof. 1. To show that G' is characteristic in G , we have to show that $f(G') = G'$ for f any automorphism of G . Now

$$f([x, y]) = f(xyx^{-1}y^{-1}) = f(x)f(y)f(x)^{-1}f(y)^{-1} = [f(x), f(y)].$$

2. We have that G' is trivial if and only if $xyx^{-1}y^{-1} = 1$ which exactly means that $xy = yx$.

3. Since G' is characteristic, it is also normal in G , and G/G' is a group. We are left to prove it is an abelian group. Take two elements (that is two cosets) $G'x$ and $G'y$ in G/G' . We have that $G'xG'y = G'yG'x \iff G'xy = G'yx$ by definition of the law group on G/G' . Now

$$G'xy = G'yx \iff xy(yx)^{-1} \in G' \iff xyx^{-1}y^{-1} \in G',$$

which holds by definition.

4. Let us assume that N is normal in G . We have that G/N is a group, and G/N is abelian if and only if for Nx, Ny two cosets we have

$$NxNy = NyNx \iff Nxy = Nyx \iff xy(yx)^{-1} \in N \iff xyx^{-1}y^{-1} \in N$$

which exactly tells that each commutator must be in N . □

We can iterate the process of taking commutators:

$$G^{(0)} = G, G^{(1)} = G', G^{(2)} = (G')', \dots, G^{(i+1)} = (G^{(i)})', \dots$$

The process may or may not reach $\{1\}$.

Definition 1.32. The group G is said to be **solvable** if $G^{(r)} = 1$ for some r . We then have a **normal series**

$$\{1\} = G^{(r)} \trianglelefteq G^{(r-1)} \trianglelefteq \dots \trianglelefteq G^{(0)} = G$$

called the **derived series** of G .

We have already seen the notion of subnormal series in the previous section. By normal series, we mean a serie where not only each group is normal in its successor, but also each group is normal in the whole group, namely each $G^{(i)}$ is normal in G . We have indeed such series here using the fact that the commutator subgroup is a characteristic subgroup, which is furthermore a transitivity property.

Let us make a few remarks about the definition of solvable group.

Lemma 1.45. 1. Every abelian group is solvable.

2. A group G both simple and solvable is cyclic of prime order.

3. A non-abelian simple group G cannot be solvable.

Proof. 1. We know that G is abelian if and only if G' is trivial. We thus get the normal series

$$G^{(0)} = G \triangleright G^{(1)} = \{1\}.$$

2. If G is simple, then its only normal subgroups are $\{1\}$ and G . Since G' is characteristic and thus normal, we have either $G' = \{1\}$ or $G' = G$. The latter cannot possibly happen, since then the derived series cannot reach $\{1\}$ which contradicts the fact that G is solvable. Thus we must have that $G' = \{1\}$, which means that G is abelian. We conclude by remembering that an abelian simple group must be cyclic of order a prime p .
3. If G is non-abelian, then G' cannot be trivial, thus since G is simple, its only normal subgroups can be either $\{1\}$ or $\{G\}$, thus G' must be either one of the other, and it cannot be $\{1\}$, so it must be G . Thus the derived series never reaches $\{1\}$ and G cannot be solvable. □

There are several ways to define solvability.

Proposition 1.46. *The following conditions are equivalent.*

1. G is solvable, that is, it has a derived series

$$\{1\} = G^{(r)} \trianglelefteq G^{(r-1)} \trianglelefteq \dots \trianglelefteq G^{(0)} = G.$$

2. G has a normal series

$$\{1\} = G_r \trianglelefteq G_{r-1} \trianglelefteq \dots \trianglelefteq G_0 = G$$

where all factors, that is all quotient groups G_i/G_{i+1} are abelian.

3. G has a subnormal series

$$\{1\} = G_r \trianglelefteq G_{r-1} \trianglelefteq \dots \trianglelefteq G_0 = G$$

where all factors, that is all quotient groups G_i/G_{i+1} are abelian.

Proof. That 1. \Rightarrow 2. is clear from Lemma 1.44 where we proved that G/G' is abelian, where G' is the commutator subgroup of G .

That 2. \Rightarrow 3. is also clear since the notion of normal series is stronger than subnormal series.

What we need to prove is thus that 3. \Rightarrow 1. Starting from G , we can always compute G' , then $G^{(2)}, \dots$. To prove that G has a derived series, we need to check that $G^{(r)} = \{1\}$. Suppose thus that G has a subnormal series

$$1 = G_r \trianglelefteq G_{r-1} \trianglelefteq \dots \trianglelefteq G_0 = G$$

where all quotient groups G_i/G_{i+1} are abelian. For $i = 0$, we get $G_1 \trianglelefteq G$ and G/G_1 is abelian. By Lemma 1.44, we know that G/G_1 is abelian is equivalent

to $G' \leq G_1$. By induction, let us assume that $G^{(i)} \leq G_i$, and see what happens with $G^{(i+1)}$. We have that $G^{(i+1)} = (G^{(i)})' \leq G'_i$ by induction hypothesis, and that $G'_i \leq G_{i+1}$ since G_i/G_{i+1} is abelian. Thus $G^{(r)} \leq G_r = \{1\}$. \square

Let us see what are the properties of subgroups and quotients of solvable groups.

Proposition 1.47. *Subgroups and quotients of a solvable group are solvable.*

Proof. Let us first consider subgroups of a solvable groups. If H is a subgroup of a solvable group G , then H is solvable because $H^{(i)} \leq G^{(i)}$ for all i , and in particular for r such that $H^{(r)} \leq G^{(r)} = \{1\}$ which proves that the derived series of H terminates.

Now consider N a normal subgroup of a solvable group G . The commutators of G/N are cosets of the form $xNyNx^{-1}Ny^{-1}N = xyx^{-1}y^{-1}N$, so that the commutator subgroup $(G/N)'$ of G/N satisfies $(G/N)' = G'N/N$ (we cannot write G'/N since there is no reason for N to be a subgroup of G'). Inductively, we have $(G/N)^{(i)} = G^{(i)}N/N$. Since G is solvable, $G^{(r)} = \{1\}$ and thus $(G/N)^{(r)} = N/N = \{1\}$ which shows that G/N is solvable. \square

Example 1.28. Consider the symmetric group S_4 . It has a subnormal series

$$\{1\} \triangleleft C_2 \times C_2 \triangleleft A_4 \triangleleft S_4,$$

where A_4 is the alternating group of order 12 (given by the even permutations of 4 elements) and $C_2 \times C_2$ is the Klein group of order 4 (corresponding to the permutations 1, (12)(34), (13)(24), (14)(23)). The quotient groups are

$$\begin{array}{lll} C_2 \times C_2 / \{1\} & \simeq C_2 \times C_2 & \text{abelian of order 4} \\ A_4 / C_2 \times C_2 & \simeq C_3 & \text{abelian of order 3} \\ S_4 / A_4 & \simeq C_2 & \text{abelian of order 2.} \end{array}$$

We finish by introducing the notion of a nilpotent group. We will skip the general definition, and consider only finite nilpotent groups, for which the following characterization is available.

Proposition 1.48. *The following statements are equivalent.*

1. G is the direct product of its Sylow subgroups.
2. Every Sylow subgroup of G is normal.

Proof. If G is the direct product of its Sylow subgroups, that every Sylow subgroup of G is normal is immediate since the factors of a direct product are normal subgroups.

Assume that every Sylow subgroup of G is normal, then by Lemma 1.35, we know that every normal Sylow p -subgroup is unique, thus there is a unique Sylow p_i -subgroup P_i for each prime divisor p_i of $|G|$, $i = 1, \dots, k$. Now by Lemma 1.30, we have that $|P_1 P_2| = |P_1| |P_2|$ since $P_1 \cap P_2 = \{1\}$, and thus

$|P_1 \cdots P_k| = |P_1| \cdots |P_k| = |G|$ by definition of Sylow subgroups. Since we work with finite groups, we deduce that G is indeed the direct product of its Sylow subgroups, having that $G = P_1 \cdots P_k$ and $P_i \cap \prod_{j \neq i} P_j$ is trivial. \square

Definition 1.33. A finite group G which is the product of its Sylow subgroups, or equivalently by the above proposition satisfies that each of its Sylow subgroup is normal is called a **nilpotent group**.

Corollary 1.49. *Every finite abelian group and every finite p -group is nilpotent.*

Proof. A finite abelian group surely has the property that each of its Sylow subgroup is normal, so it is nilpotent.

Now consider P a finite p -group. Then by definition P has only one Sylow subgroup, namely itself, so it is the direct product of its Sylow subgroups and thus is nilpotent. \square

Definition 1.34. A **central series** for a group G is a normal series

$$1 = G_0 \trianglelefteq G_1 \trianglelefteq \cdots \trianglelefteq G_r = G$$

such that $G_i/G_{i-1} \subseteq Z(G/G_{i-1})$ for every $i = 1, \dots, r$. An arbitrary group G is said to be **nilpotent** if it has a central series.

Nilpotent groups in general are discussed with solvable groups since they can be described with normal series, and one can prove that they are solvable. Indeed, if $G_i/G_{i-1} \subseteq Z(G/G_{i-1})$, then the elements of G_i/G_{i-1} commute with each other, since they commute with everything in G/G_{i-1} , thus G_i/G_{i-1} is abelian.

Proposition 1.50. *If G is a finite nilpotent group, then no proper subgroup H of G is equal to its normalizer $N_G(H) = \{g \in G, gH = Hg\}$.*

Proof. Let H be a proper subgroup of G , and let n be the largest index such that $G_n \subseteq H$ (such index exists since G is nilpotent). There exists $a \in G_{n+1}$ such that $a \notin H$ (since H is a proper subgroup). Now for every $h \in H$, the cosets aG_n and hG_n commute (since $G_{n+1}/G_n \subseteq Z(G/G_n)$), namely:

$$G_n ah = (G_n a)(G_n h) = (G_n h)(G_n a) = G_n ha$$

and thus there is some $h' \in G_n \subseteq H$ for which

$$ah = h'ha$$

that is

$$aha^{-1} = h'h \in H.$$

Thus $a \in N_G(H)$ and $a \notin H$. \square

The main definitions and results of this chapter are

- **(1.1-1.2).** Definitions of: group, subgroup, group homomorphism, order of a group, order of an element, cyclic group.
- **(1.3-1.4).** Lagrange's Theorem. Definitions of: coset, normal subgroup, quotient group
- **(1.5).** 1st, 2nd and 3rd Isomorphism Theorems.
- **(1.6).** Definitions of: external (semi-)direct product, internal (semi-)direct product.
- **(1.7).** Cayley's Theorem, the Orbit-Stabilizer Theorem, the Orbit-Counting Theorem. Definitions of: symmetric group, group action, orbit, transitive action, stabilizer, centralizer. That the orbits partition the set under the action of a group
- **(1.8).** Definition: p -group, Sylow p -subgroup. The 3 Sylow Theorems, Cauchy Theorem
- **(1.9).** Definition: simple group. Applications of the Sylow Theorems.
- **(1.10).** Definitions: subnormal series, composition series. Jordan-Hölder Theorem.
- **(1.11).** Definitions: characteristic subgroup, commutator subgroup, normal and derived series, solvable group, finite nilpotent group.

Chapter 2

Ring Theory

In the first section below, a ring will be defined as an abstract structure with a commutative addition, and a multiplication which may or may not be commutative. This distinction yields two quite different theories: the theory of respectively commutative or non-commutative rings. These notes are mainly concerned about commutative rings.

Non-commutative rings have been an object of systematic study only quite recently, during the 20th century. Commutative rings on the contrary have appeared though in a hidden way much before, and as many theories, it all goes back to Fermat's Last Theorem.

In 1847, the mathematician Lamé announced a solution of Fermat's Last Theorem, but Liouville noticed that the proof depended on a unique decomposition into primes, which he thought was unlikely to be true. Though Cauchy supported Lamé, Kummer was the one who finally published an example in 1844 to show that the uniqueness of prime decompositions failed. Two years later, he restored the uniqueness by introducing what he called "ideal complex numbers" (today, simply "ideals") and used it to prove Fermat's Last Theorem for all $n < 100$ except $n = 37, 59, 67$ and 74 .

It is Dedekind who extracted the important properties of "ideal numbers", defined an "ideal" by its modern properties: namely that of being a subgroup which is closed under multiplication by any ring element. He further introduced prime ideals as a generalization of prime numbers. Note that today we still use the terminology "Dedekind rings" to describe rings which have in particular a good behavior with respect to factorization of prime ideals. In 1882, an important paper by Dedekind and Weber developed the theory of rings of polynomials. At this stage, both rings of polynomials and rings of numbers (rings appearing in the context of Fermat's Last Theorem, such as what we call now the Gaussian integers) were being studied. But it was separately, and no one made connection between these two topics. Dedekind also introduced the term "field" (Körper) for a commutative ring in which every non-zero element has a

multiplicative inverse but the word “ring” is due to Hilbert, who, motivated by studying invariant theory, studied ideals in polynomial rings proving his famous “Basis Theorem” in 1893.

It will take another 30 years and the work of Emmy Noether and Krull to see the development of axioms for rings. Emmy Noether, about 1921, is the one who made the important step of bringing the two theories of rings of polynomials and rings of numbers under a single theory of abstract commutative rings.

In contrast to commutative ring theory, which grew from number theory, non-commutative ring theory developed from an idea of Hamilton, who attempted to generalize the complex numbers as a two dimensional algebra over the reals to a three dimensional algebra. Hamilton, who introduced the idea of a vector space, found inspiration in 1843, when he understood that the generalization was not to three dimensions but to four dimensions and that the price to pay was to give up the commutativity of multiplication. The quaternion algebra, as Hamilton called it, launched non-commutative ring theory.

Other natural non-commutative objects that arise are matrices. They were introduced by Cayley in 1850, together with their laws of addition and multiplication and, in 1870, Pierce noted that the now familiar ring axioms held for square matrices.

An early contributor to the theory of non-commutative rings was the Scottish mathematician Wedderburn, who in 1905, proved “Wedderburn’s Theorem”, namely that every finite division ring is commutative and so is a field.

It is only around the 1930’s that the theories of commutative and non-commutative rings came together and that their ideas began to influence each other.

2.1 Rings, ideals and homomorphisms

Definition 2.1. A **ring** R is an abelian group with a multiplication operation

$$(a, b) \mapsto ab$$

which is associative, and satisfies the distributive laws

$$a(b + c) = ab + ac, (a + b)c = ac + bc$$

with identity element 1.

There is a group structure with the addition operation, but not necessarily with the multiplication operation. Thus an element of a ring may or may not be invertible with respect to the multiplication operation. Here is the terminology used.

Definition 2.2. Let a, b be in a ring R . If $a \neq 0$ and $b \neq 0$ but $ab = 0$, then we say that a and b are **zero divisors**. If $ab = ba = 1$, we say that a is a **unit** or that a is **invertible**.

While the addition operation is commutative, it may or not be the case with the multiplication operation.

Definition 2.3. Let R be ring. If $ab = ba$ for any a, b in R , then R is said to be **commutative**.

Here are the definitions of two particular kinds of rings where the multiplication operation behaves well.

Definition 2.4. An **integral domain** is a commutative ring with no zero divisor. A **division ring** or **skew field** is a ring in which every non-zero element a has an inverse a^{-1} .

Let us give two more definitions and then we will discuss several examples.

Definition 2.5. The **characteristic** of a ring R , denoted by $\text{char}R$, is the smallest positive integer such that

$$n \cdot 1 = \underbrace{1 + 1 + \dots + 1}_{n \text{ times}} = 0.$$

We can also extract smaller rings from a given ring.

Definition 2.6. A **subring** of a ring R is a subset S of R that forms a ring under the operations of addition and multiplication defined in R .

Examples 2.1. 1. \mathbb{Z} is an integral domain but not a field.

2. The integers modulo n form a ring, which is an integral domain if and only if n is prime.
3. The $n \times n$ matrices $\mathcal{M}_n(\mathbb{R})$ with coefficients in \mathbb{R} are a ring, but not an integral domain if $n \geq 2$.
4. Let us construct the smallest and also most famous example of division ring. Take $1, i, j, k$ to be basis vectors for a 4-dimensional vector space over \mathbb{R} , and define multiplication by

$$i^2 = j^2 = k^2 = -1, \quad ij = k, \quad jk = i, \quad ki = j, \quad ji = -ij, \quad kj = -jk, \quad ik = -ki.$$

Then

$$\mathbb{H} = \{a + bi + cj + dk, \quad a, b, c, d \in \mathbb{R}\}$$

forms a division ring, called the **Hamilton's quaternions**. So far, we have only seen the ring structure. Let us now discuss the fact that every non-zero element is invertible. Define the **conjugate** of an element $h = a + bi + cj + dk \in \mathbb{H}$ to be $\bar{h} = a - bi - cj - dk$ (yes, exactly the same way you did it for complex numbers). It is an easy computation (and a good exercise if you are not used to the non-commutative world) to check that

$$q\bar{q} = a^2 + b^2 + c^2 + d^2.$$

Now take q^{-1} to be

$$q^{-1} = \frac{\bar{q}}{q\bar{q}}.$$

Clearly $qq^{-1} = q^{-1}q = 1$ and the denominator cannot possibly be 0, but if $a = b = c = d = 0$.

5. If R is a ring, then the set $R[X]$ of polynomials with coefficients in R is a ring.

Similarly to what we did with groups, we now define a map from a ring to another which has the property of carrying one ring structure to the other.

Definition 2.7. Let R, S be two rings. A map $f : R \rightarrow S$ satisfying

1. $f(a + b) = f(a) + f(b)$ (this is thus a group homomorphism)
2. $f(ab) = f(a)f(b)$
3. $f(1_R) = 1_S$

for $a, b \in R$ is called **ring homomorphism**.

The notion of “ideal number” was introduced by the mathematician Kummer, as being some special “numbers” (well, nowadays we call them groups) having the property of unique factorization, even when considered over more general rings than \mathbb{Z} (a bit of algebraic number theory would be good to make this more precise). Today only the name “ideal” is left, and here is what it gives in modern terminology:

Definition 2.8. Let \mathcal{I} be a subset of a ring R . Then an additive subgroup of R having the property that

$$ra \in \mathcal{I} \text{ for } a \in \mathcal{I}, r \in R$$

is called a **left ideal** of R . If instead we have

$$ar \in \mathcal{I} \text{ for } a \in \mathcal{I}, r \in R$$

we say that we have a **right ideal** of R . If an ideal happens to be both a right and a left ideal, then we call it a **two-sided ideal** of R , or simply an ideal of R .

Of course, for any ring R , both R and $\{0\}$ are ideals. We thus introduce some terminology to precise whether we consider these two trivial ideals.

Definition 2.9. We say that an ideal \mathcal{I} of R is **proper** if $\mathcal{I} \neq R$. We say that is it **non-trivial** if $\mathcal{I} \neq R$ and $\mathcal{I} \neq 0$.

If $f : R \rightarrow S$ is a ring homomorphism, we define the kernel of f in the most natural way:

$$\text{Ker } f = \{r \in R, f(r) = 0\}.$$

Since a ring homomorphism is in particular a group homomorphism, we already know that f is injective if and only if $\text{Ker } f = \{0\}$. It is easy to check that $\text{Ker } f$ is a proper two-sided ideal:

- $\text{Ker } f$ is an additive subgroup of R .
- Take $a \in \text{Ker } f$ and $r \in R$. Then

$$f(ra) = f(r)f(a) = 0 \text{ and } f(ar) = f(a)f(r) = 0$$

showing that ra and ar are in $\text{Ker } f$.

- Then $\text{Ker } f$ has to be proper (that is, $\text{Ker } f \neq R$), since $f(1) = 1$ by definition.

We can thus deduce the following (extremely useful) result.

Lemma 2.1. *Suppose $f : R \rightarrow S$ is a ring homomorphism and the only two-sided ideals of R are $\{0\}$ and R . Then f is injective.*

Proof. Since $\text{Ker } f$ is a two-sided ideal of R , then either $\text{Ker } f = \{0\}$ or $\text{Ker } f = R$. But $\text{Ker } f \neq R$ since $f(1) = 1$ by definition (in words, $\text{Ker } f$ is a proper ideal). \square

At this point, it may be worth already noticing the analogy between on the one hand rings and their two-sided ideals, and on the other hand groups and their normal subgroups.

- Two-sided ideals are stable when the ring acts on them by multiplication, either on the right or on the left, and thus

$$rar^{-1} \in \mathcal{I}, \quad a \in \mathcal{I}, \quad r \in R,$$

while normal subgroups are stable when the groups on them by conjugation

$$ghg^{-1} \in H, \quad h \in H, \quad g \in G \quad (H \leq G).$$

- Groups with only trivial normal subgroups are called simple. We will not see it formally here, but rings with only trivial two-sided ideals as in the above lemma are called simple rings.
- The kernel of a group homomorphism is a normal subgroup, while the kernel of a ring homomorphism is an ideal.
- Normal subgroups allowed us to define quotient groups. We will see now that two-sided ideals will allow to define quotient rings.

2.2 Quotient rings

Let \mathcal{I} be a proper two-sided ideal of R . Since \mathcal{I} is an additive subgroup of R by definition, it makes sense to speak of cosets $r + \mathcal{I}$ of \mathcal{I} , $r \in R$. Furthermore, a ring has a structure of abelian group for addition, so \mathcal{I} satisfies the definition of a normal subgroup. From group theory, we thus know that it makes sense to speak of the quotient group

$$R/\mathcal{I} = \{r + \mathcal{I}, \quad r \in R\},$$

group which is actually abelian (inherited from R being an abelian group for the addition).

We now endow R/\mathcal{I} with a multiplication operation as follows. Define

$$(r + \mathcal{I})(s + \mathcal{I}) = rs + \mathcal{I}.$$

Let us make sure that this is well-defined, namely that it does not depend on the choice of the representative in each coset. Suppose that

$$r + \mathcal{I} = r' + \mathcal{I}, \quad s + \mathcal{I} = s' + \mathcal{I},$$

so that $a = r' - r \in \mathcal{I}$ and $b = s' - s \in \mathcal{I}$. Now

$$r's' = (a + r)(b + s) = ab + as + rb + rs \in rs + \mathcal{I}$$

since ab, as and rb belongs to \mathcal{I} using that $a, b \in \mathcal{I}$ and the definition of ideal. This tells us $r's'$ is also in the coset $rs + \mathcal{I}$ and thus multiplication does not depend on the choice of representatives. Note though that this is true only because we assumed a two-sided ideal \mathcal{I} , otherwise we could not have concluded, since we had to deduce that both as and rb are in \mathcal{I} .

Definition 2.10. The set of cosets of the two-sided ideal \mathcal{I} given by

$$R/\mathcal{I} = \{r + \mathcal{I}, r \in R\}$$

is a ring with identity $1_R + \mathcal{I}$ and zero element $0_R + \mathcal{I}$ called a **quotient ring**.

Note that we need the assumption that \mathcal{I} is a proper ideal of R to claim that R/\mathcal{I} contains both an identity and a zero element (if $R = \mathcal{I}$, then R/\mathcal{I} has only one element).

Example 2.2. Consider the ring of matrices $\mathcal{M}_2(\mathbb{F}_2[i])$, where \mathbb{F}_2 denotes the integers modulo 2, and i is such that $i^2 = -1 \equiv 1 \pmod{2}$. This is thus the ring of 2×2 matrices with coefficients in

$$\mathbb{F}_2[i] = \{a + ib, a, b \in \{0, 1\}\}.$$

Let \mathcal{I} be the subset of matrices with coefficients taking values 0 and $1 + i$ only. It is a two-sided ideal of $\mathcal{M}_2(\mathbb{F}_2[i])$. Indeed, take a matrix $U \in \mathcal{I}$, a matrix $M \in \mathcal{M}_2(\mathbb{F}_2[i])$, and compute UM and MU . An immediate computation shows that all coefficients are of the form $a(1 + i)$ with $a \in \mathbb{F}_2[i]$, that is all coefficients are in $\{0, 1 + i\}$. Clearly \mathcal{I} is an additive group.

We then have a quotient ring

$$\mathcal{M}_2(\mathbb{F}_2[i])/\mathcal{I}.$$

We have seen that $\text{Ker } f$ is a proper two-sided ideal when f is a ring homomorphism. We now prove the converse.

Proposition 2.2. *Every proper two-sided ideal \mathcal{I} is the kernel of a ring homomorphism.*

Proof. Consider the canonical projection π that we know from group theory. Namely

$$\pi : R \rightarrow R/\mathcal{I}, \quad r \mapsto \pi(r) = r + \mathcal{I}.$$

We already know that π is group homomorphism, and that its kernel is \mathcal{I} . We are only left to prove that π is a ring homomorphism:

- since \mathcal{I} is two-sided, then R/\mathcal{I} is a ring.
- $\pi(rs) = rs + \mathcal{I} = (r + \mathcal{I})(s + \mathcal{I}) = \pi(r)\pi(s)$.
- $\pi(1_R) = 1_R + \mathcal{I}$ which is indeed the identity element of R/\mathcal{I} .

□

We are now ready to state a factor theorem and a 1st isomorphism theorem for rings, the same way we did for groups. It may help to keep in mind the analogy between two-sided ideals and normal subgroups mentioned above.

Assume that we have a ring R which contains a proper two-sided ideal \mathcal{I} , another ring S , and $f : R \rightarrow S$ a ring homomorphism. Let π be the canonical projection from R to the quotient group R/\mathcal{I} :

$$\begin{array}{ccc} R & \xrightarrow{f} & S \\ \pi \downarrow & \nearrow \bar{f} & \\ R/\mathcal{I} & & \end{array}$$

We would like to find a ring homomorphism $\bar{f} : R/\mathcal{I} \rightarrow S$ that makes the diagram commute, namely

$$f(a) = \bar{f}(\pi(a))$$

for all $a \in R$.

Theorem 2.3. (Factor Theorem for Rings). *Any ring homomorphism f whose kernel K contains \mathcal{I} can be factored through R/\mathcal{I} . In other words, there is a unique ring homomorphism $\bar{f} : R/\mathcal{I} \rightarrow S$ such that $\bar{f} \circ \pi = f$. Furthermore*

1. \bar{f} is an epimorphism if and only if f is.
2. \bar{f} is a monomorphism if and only if $K = \mathcal{I}$.
3. \bar{f} is an isomorphism if and only if f is an epimorphism and $K = \mathcal{I}$.

Proof. Since we have already done the proof for groups with many details, here we will just mention a few important points in the proof.

Let $a + \mathcal{I} \in R/\mathcal{I}$ such that $\pi(a) = a + \mathcal{I}$ for $a \in R$. We define

$$\bar{f}(a + \mathcal{I}) = f(a).$$

This is the most natural way to do it, however, we need to make sure that this is indeed well-defined, in the sense that it should not depend on the choice of

the representative taken in the coset. Let us thus take another representative, say $b \in a + \mathcal{I}$. Since a and b are in the same coset, they satisfy $a - b \in \mathcal{I} \subset K$, where $K = \text{Ker}(f)$ by assumption. Since $a - b \in K$, we have $f(a - b) = 0$ and thus $f(a) = f(b)$.

Now that \bar{f} is well defined, it is an easy computation to check that \bar{f} inherits the property of ring homomorphism from f .

The rest of the proof works exactly the same as for groups. \square

The first isomorphism theorem for rings is similar to the one for groups.

Theorem 2.4. (1st Isomorphism Theorem for Rings). *If $f : R \rightarrow S$ is a ring homomorphism with kernel K , then the image of f is isomorphic to R/K :*

$$\text{Im}(f) \simeq R/\text{Ker}(f).$$

Proof. We know from the Factor Theorem that

$$\bar{f} : R/\text{Ker}(f) \rightarrow S$$

is an isomorphism if and only if f is an epimorphism, and clearly f is an epimorphism on its image, which concludes the proof. \square

Example 2.3. Let us finish Example 2.2. We showed there that $\mathcal{M}_2(\mathbb{F}_2[i])/\mathcal{I}$ is a quotient ring, where \mathcal{I} is the ideal formed of matrices with coefficients in $\{0, 1 + i\}$. Consider the ring homomorphism:

$$f : \mathcal{M}_2(\mathbb{F}_2[i]) \rightarrow \mathcal{M}_2(\mathbb{F}_2), M = (m_{k,l}) \mapsto f(M) = (m_{k,l} \pmod{1+i})$$

that is f looks at the coefficients of $M \pmod{1+i}$. Its kernel is \mathcal{I} and it is surjective. By the first isomorphism for rings, we have

$$\mathcal{M}_2(\mathbb{F}_2[i])/\mathcal{I} \simeq \mathcal{M}_2(\mathbb{F}_2).$$

2.3 The Chinese Remainder Theorem

We will prove a “general” Chinese Remainder Theorem, rephrased in terms of rings and ideals.

For that let us start by introducing some new definitions about ideals, that will collect some of the manipulations one can do on ideals. Let us start with the sum.

Definition 2.11. Let \mathcal{I} and \mathcal{J} be two ideals of a ring R . The **sum** of \mathcal{I} and \mathcal{J} is the ideal

$$\mathcal{I} + \mathcal{J} = \{x + y, x \in \mathcal{I}, y \in \mathcal{J}\}.$$

If \mathcal{I} and \mathcal{J} are right (resp. left) ideals, so is their sum.

Note that the intersection $\mathcal{I} \cap \mathcal{J}$ of two (resp. right, left, two-sided) ideals of R is again a (resp. right, left, two-sided) ideal of R . We can define a notion of being co-prime for ideals as follows.

Definition 2.12. The ideals \mathcal{I} and \mathcal{J} of R a commutative ring are **relatively prime** if

$$\mathcal{I} + \mathcal{J} = R.$$

Notice that if \mathcal{I} and \mathcal{J} are relatively prime then

$$\mathcal{I}\mathcal{J} = \mathcal{I} \cap \mathcal{J}.$$

Indeed, we clearly have that

$$\mathcal{I}\mathcal{J} \subset \mathcal{I} \cap \mathcal{J}$$

since $\mathcal{I}\mathcal{J}$ contains by definition sums of elements xy , $x \in \mathcal{I}, y \in \mathcal{J}$, with $xy \in \mathcal{I}$ and $xy \in \mathcal{J}$ by definition of two-sided ideal. Conversely

$$\mathcal{I} \cap \mathcal{J} \subset \mathcal{I}\mathcal{J}$$

since there exist $x \in \mathcal{I}, y \in \mathcal{J}$ such that $x + y = 1$ by definition of relatively prime, and for every element $a \in \mathcal{I} \cap \mathcal{J}$, we have that

$$a = a(x + y) = ax + ay \in \mathcal{I}\mathcal{J}.$$

Finally, let us extend the notion of “modulo” to ideals.

Definition 2.13. If $a, b \in R$ and \mathcal{I} is an ideal of R , we say that a is **congruent to b modulo \mathcal{I}** if

$$a - b \in \mathcal{I}.$$

A last definition this time about rings is needed before we can state the theorem.

Definition 2.14. If R_1, \dots, R_n are rings, the **direct product** of the R_i is defined as the ring of n -tuples (a_1, \dots, a_n) , $a_i \in R_i$, with componentwise addition and multiplication. The zero element is $(0, \dots, 0)$ and the identity is $(1, \dots, 1)$ where 1 means 1_{R_i} for each i .

This definition is an immediate generalization of the direct product we studied for groups.

Theorem 2.5. (Chinese Remainder Theorem). *Let R be a commutative ring, and let $\mathcal{I}_1, \dots, \mathcal{I}_n$ be ideals in R , such that*

$$\mathcal{I}_i + \mathcal{I}_j = R, \quad i \neq j.$$

1. *If a_1, \dots, a_n are elements of R , there exists an element $a \in R$ such that*

$$a \equiv a_i \pmod{\mathcal{I}_i}, \quad i = 1, \dots, n.$$

2. *If b is another element of R such that $b \equiv a_i \pmod{\mathcal{I}_i}, i = 1, \dots, n$, then*

$$b \equiv a \pmod{\cap_{i=1}^n \mathcal{I}_i}.$$

Conversely, if b satisfies the above congruence, then $b \equiv a_i \pmod{\mathcal{I}_i}, i = 1, \dots, n$.

3. We have that

$$R/\bigcap_{i=1}^n \mathcal{I}_i \simeq \prod_{i=1}^n R/\mathcal{I}_i.$$

Proof. 1. For $j > 1$, we have by assumption that $\mathcal{I}_1 + \mathcal{I}_j = R$, and thus there exist $b_j \in \mathcal{I}_1$ and $d_j \in \mathcal{I}_j$ such that

$$b_j + d_j = 1, \quad j = 2, \dots, n.$$

This yields that

$$\prod_{j=2}^n (b_j + d_j) = 1. \quad (2.1)$$

Now if we look at the left hand side of the above equation, we have

$$(b_2 + d_2)(b_3 + d_3) \cdots (b_n + d_n) = \underbrace{(b_2 b_3 + b_2 d_3 + d_2 b_3 + d_2 d_3)}_{\in \mathcal{I}_1} \cdots (b_n + d_n)$$

and all the terms actually belong to \mathcal{I}_1 , but $c_1 := \prod_{j=2}^n d_j \in \prod_{j=2}^n \mathcal{I}_j$. Thus

$$c_1 \equiv 1 \pmod{\mathcal{I}_1}$$

from (2.1). On the other hand, we also have

$$c_1 \equiv 0 \pmod{\mathcal{I}_j}$$

for $j > 1$ since $c_1 \in \prod_{j=2}^n \mathcal{I}_j$.

More generally, for all i , we can find c_i with

$$c_i \equiv 1 \pmod{\mathcal{I}_i}, \quad c_i \equiv 0 \pmod{\mathcal{I}_j}, \quad j \neq i.$$

Now take arbitrary elements $a_1, \dots, a_n \in R$, and set

$$a = a_1 c_1 + \dots + a_n c_n.$$

Let us check that a is the solution we are looking for. Rewrite

$$a - a_i = a - a_i c_i + a_i c_i - a_i = (a - a_i c_i) + a_i (c_i - 1).$$

If we look modulo \mathcal{I}_i , we get

$$a - a_i \equiv a - a_i c_i \equiv a_1 + \dots + a_{i-1} c_{i-1} + a_{i+1} c_{i+1} + \dots + a_n c_n \equiv 0 \pmod{\mathcal{I}_i}$$

where the first congruence follows from $c_i - 1 \equiv 0 \pmod{\mathcal{I}_i}$ and the third congruence comes from $c_j \equiv 0 \pmod{\mathcal{I}_j}, j \neq i$.

2. We have just shown the existence of a solution a modulo \mathcal{I}_i for $i = 1, \dots, n$. We now discuss the question of unicity, and show that the solution is actually not unique, but any other solution than a is actually congruent to $a \pmod{\cap_{i=1}^n \mathcal{I}_i}$.

We have for all $i = 1, \dots, n$ that

$$b \equiv a_i \pmod{\mathcal{I}_i} \iff b \equiv a \pmod{\mathcal{I}_i} \iff b - a \equiv 0 \pmod{\mathcal{I}_i}$$

which finally is equivalent to

$$b - a \in \cap_{i=1}^n \mathcal{I}_i.$$

3. Define the ring homomorphism $f : R \rightarrow \prod_{i=1}^n R/\mathcal{I}_i$, sending

$$a \mapsto f(a) = (a + \mathcal{I}_1, \dots, a + \mathcal{I}_n).$$

- This map is surjective: take any $(a_1 + \mathcal{I}_1, \dots, a_n + \mathcal{I}_n) \in \prod_{i=1}^n R/\mathcal{I}_i$, then we must find an $a \in R$ such that $f(a) = (a_1 + \mathcal{I}_1, \dots, a_n + \mathcal{I}_n)$, that is $a + \mathcal{I}_i = a_i + \mathcal{I}_i$, or equivalently $a_i \equiv a \pmod{\mathcal{I}_i}$, which is true by the first point.
- Its kernel is given by

$$\begin{aligned} \text{Ker } f &= \{a \in R, f(a) = (\mathcal{I}_1, \dots, \mathcal{I}_n)\} \\ &= \{a \in R, a \in \mathcal{I}_i, i = 1, \dots, n\} \\ &= \prod_{i=1}^n \mathcal{I}_i. \end{aligned}$$

We conclude using the first isomorphism Theorem for rings. □

The Chinese remainder Theorem does not hold in the non-commutative case. Consider the ring R of non-commutative real polynomials in X and Y . Denote by \mathcal{I} the principal two-sided ideal generated by X and \mathcal{J} the principal two-sided ideal generated by $XY + 1$. Then $\mathcal{I} + \mathcal{J} = R$ but $\mathcal{I} \cap \mathcal{J} \neq \mathcal{I}\mathcal{J}$. Recall that we showed after Definition 2.12 that in the commutative case, if \mathcal{I} and \mathcal{J} are relatively prime, then $\mathcal{I}\mathcal{J} = \mathcal{I} \cap \mathcal{J}$.

2.4 Maximal and prime ideals

Here are a few special ideals.

Definition 2.15. The **ideal generated** by the non-empty set X of R is the smallest ideal of R that contains X . It is denoted by $\langle X \rangle$. It is the collection of all finite sums of the form $\sum_i r_i x_i s_i$.

Definition 2.16. An ideal generated by a single element a is called a **principal ideal**, denoted by $\langle a \rangle$.

Definition 2.17. A **maximal ideal** in the ring R is a proper ideal that is not contained in any strictly larger proper ideal.

One can prove that every proper ideal is contained in a maximal ideal, and that consequently every ring has at least one maximal ideal. We skip the proof here, since it heavily relies on set theory, requires many new definitions and the use of Zorn's lemma.

Instead, let us mention that a correspondence Theorem exists for rings, the same way it exists for groups, since we will need it for characterizing maximal ideals.

Theorem 2.6. (Correspondence Theorem for rings). *If \mathcal{I} is an ideal of a ring R , then the canonical map*

$$\pi : R \rightarrow R/\mathcal{I}$$

sets up a one-to-one correspondence between

- *the set of all subrings of R containing \mathcal{I} and the set of all subrings of R/\mathcal{I} ,*
- *the set of all ideals of R containing \mathcal{I} and the set of all ideals of R/\mathcal{I} .*

Here is a characterization of maximal ideals in commutative rings.

Theorem 2.7. *Let M be an ideal in the commutative ring R . We have*

$$M \text{ maximal} \iff R/M \text{ is a field.}$$

Proof. Let us start by assuming that M is maximal. Since R/M is a ring, we need to find the multiplicative inverse of $a+M \in R/M$ assuming that $a+M \neq 0$ in R/M , that is $a \notin M$. Since M is maximal, the ideal $Ra + M$ has to be R itself, since $M \subset Ra + M$. Thus $1 \in Ra + M = R$, that is

$$1 = ra + m, \quad r \in R, \quad m \in M.$$

Then

$$(r + M)(a + M) = ra + M = (1 - m) + M = 1 + M$$

proving that $r + M$ is $(a + M)^{-1}$.

Conversely, let us assume that R/M is a field. First we notice that M must be a proper ideal of R , since if $M = R$, then R/M contains only one element and $1 = 0$.

Let N be an ideal of R such that $M \subset N \subset R$ and $N \neq R$. We have to prove that $M = N$ to conclude that M is maximal.

By the correspondence Theorem for rings, we have a one-to-one correspondence between the set of ideals of R containing M , and the set of ideals of R/M . Since N is such an ideal, its image $\pi(N) \in R/M$ must be an ideal of R/M , and thus must be either $\{0\}$ or R/M . The latter yields that $N = R$, which is a contradiction, letting as only possibility that $\pi(N) = \{0\}$, and thus $N = M$, which completes the proof. \square

Definition 2.18. A **prime ideal** in a commutative ring R is a proper ideal P of R such that for any $a, b \in R$, we have that

$$ab \in P \Rightarrow a \in P \text{ or } b \in P.$$

Here is again a characterization of a prime ideal P of R in terms of its quotient ring R/P .

Theorem 2.8. *If P is an ideal in the commutative ring R*

$$P \text{ is a prime ideal} \iff R/P \text{ is an integral domain.}$$

Proof. Let us start by assuming that P is prime. It is thus proper by definition, and R/P is a ring. We must show that the definition of integral domain holds, namely that

$$(a + P)(b + P) = 0 + P \Rightarrow a + P = P \text{ or } b + P = P.$$

Since

$$(a + P)(b + P) = ab + P = 0 + P,$$

we must have $ab \in P$, and thus since P is prime, either $a \in P$ or $b \in P$, implying respectively that either $a + P = P$ or $b + P = P$.

Conversely, if R/P is an integral domain, then P must be proper (otherwise $1 = 0$). We now need to check the definition of a prime ideal. Let us thus consider $ab \in P$, implying that

$$(a + P)(b + P) = ab + P = 0 + P.$$

Since R/P is an integral domain, either $a + P = P$ or $b + P = P$, that is

$$a \in P \text{ or } b \in P,$$

which concludes the proof. □

Corollary 2.9. *In a commutative ring, a maximal ideal is prime.*

Proof. If M is maximal, then R/M is a field, and thus an integral domain, so that M is prime. □

Corollary 2.10. *Let $f : R \rightarrow S$ be an epimorphism of commutative rings.*

1. *If S is a field, then $\text{Ker } f$ is a maximal ideal of R .*
2. *If S is an integral domain, then $\text{Ker } f$ is a prime ideal of R .*

Proof. By the first isomorphism theorem for rings, we have that

$$S \simeq R/\text{Ker } f.$$

□

Example 2.4. Consider the ring $\mathbb{Z}[X]$ of polynomials with coefficients in \mathbb{Z} , and the ideal generated by the indeterminate X , that is $\langle X \rangle$ is the set of polynomials with constant coefficient 0. Clearly $\langle X \rangle$ is a proper ideal. To show that it is prime, consider the following ring homomorphism:

$$\varphi : \mathbb{Z}[X] \rightarrow \mathbb{Z}, f(X) \mapsto \varphi(f(X)) = f(0).$$

We have that $\langle X \rangle = \text{Ker}\varphi$ which is prime by the above corollary.

2.5 Polynomial rings

For this section, we assume that R is a commutative ring. Set $R[X]$ to be the set of polynomials in the indeterminate X with coefficients in R . It is easy to see that $R[X]$ inherits the properties of ring from R .

We define the **evaluation map** E_x , which evaluates a polynomial $f(X) \in R[X]$ in $x \in R$, as

$$E_x : R[X] \rightarrow R, f(X) \mapsto f(X)|_{X=x} = f(x).$$

We can check that E_x is a ring homomorphism.

The **degree** of a polynomial is defined as usual, that is, if $p(X) = a_0 + a_1X + \dots + a_nX^n$ with $a_n \neq 0$, then $\deg(p(X)) = \deg p = n$. By convention, we set $\deg(0) = -\infty$.

Euclidean division will play an important role in what will follow. Let us start by noticing that there exists a polynomial division algorithm over $R[X]$, namely: if $f, g \in R[X]$, with g monic, then there exist unique polynomials q and r in $R[X]$ such that

$$f = qg + r, \quad \deg r < \deg g.$$

The requirement that g is monic comes from R being a ring and not necessarily a field. If R is a field, g does not have to be monic, since one can always multiply g by the inverse of the leading coefficient, which is not possible if R is not a field.

This gives the following:

Theorem 2.11. (Remainder Theorem). *If $f \in R[X]$, $a \in R$, then there exists a unique polynomial $q(X) \in R[X]$ such that*

$$f(X) = q(X)(X - a) + f(a).$$

Hence $f(a) = 0 \iff X - a \mid f(X)$.

Proof. Since $(X - a)$ is monic, we can do the division

$$f(X) = q(X)(X - a) + r(X).$$

But now since $\deg r < \deg(X - a)$, $r(X)$ must be a constant polynomial, which implies that

$$f(a) = r(X)$$

and thus

$$f(X) = q(X)(X - a) + f(a)$$

as claimed. Furthermore, we clearly have that

$$f(a) = 0 \iff X - a \mid f(X).$$

□

The following result sounds well known, care should be taken not to generalize it to rings which are not integral domain!

Theorem 2.12. *If R is an integral domain, then a non-zero polynomial f in $R[X]$ of degree n has at most n roots in R , counting multiplicity.*

Proof. If f has no root in $R[X]$, then we are done. Let us thus assume that f has a root a_1 in R , that is $f(a_1) = 0$. Then

$$X - a_1 \mid f(X)$$

by the remainder Theorem above, meaning that

$$f(X) = q_1(X)(X - a_1)^{n_1}$$

where $q_1(a_1) \neq 0$ and $\deg q_1 = n - n_1$ since R is an integral domain. Now if a_1 is the only root of f in R , then $n_1 \leq n$ and we are done. If not, consider similarly $a_2 \neq a_1$ another root of f , so that

$$0 = f(a_2) = q_1(a_2)(a_2 - a_1)^{n_1}.$$

Since R is an integral domain, we must have that $q_1(a_2) = 0$, and thus a_2 is a root of $q_1(X)$. We can repeat the process with $q_1(X)$ instead of $f(X)$: since a_2 is a root of $q_1(X)$, we have

$$q_1(X) = q_2(X)(X - a_2)^{n_2}$$

with $q_2(a_2) \neq 0$ and $\deg q_2 = n - n_1 - n_2$. By going on iterating the process, we obtain

$$\begin{aligned} f(X) &= q_1(X)(X - a_1)^{n_1} \\ &= q_2(X)(X - a_2)^{n_2}(X - a_1)^{n_1} \\ &= \dots \\ &= (X - a_1)^{n_1}(X - a_2)^{n_2} \dots (X - a_k)^{n_k} \cdot c(X) \end{aligned}$$

where $c(X)$ is a polynomial with no root in R , possibly constant, and

$$n \geq n_1 + n_2 + \dots + n_k.$$

Since R is an integral domain, the only possible roots of f are a_1, \dots, a_k , $k \leq n$, and the number of roots counting multiplicity is less than n . □

Example 2.5. Take $R = \mathbb{Z}_8$ the ring of integers modulo 8. Consider the polynomial

$$f(X) = X^3.$$

It is easy to check that it has 4 roots: 0, 2, 4, 6. This comes from the fact that \mathbb{Z}_8 is not an integral domain.

2.6 Unique factorization and Euclidean division

In this section, all rings are assumed to be integral domains.

Let us start by defining formally the notions of irreducible and prime. The elements a, b, c, u in the definitions below all belong to an integral domain R .

Definition 2.19. The elements a, b are called **associate** if $a = ub$ for some unit u .

Definition 2.20. Let a be a non-zero element which is not a unit. Then a is said to be **irreducible** if $a = bc$ implies that either b or c must be a unit.

Definition 2.21. Let a be a non-zero element which is not a unit. Then a is called **prime** if whenever $a \mid bc$, then $a \mid b$ or $a \mid c$.

Between prime and irreducible, which notion is the stronger? The answer is in the proposition below.

Proposition 2.13. *If a is prime, then a is irreducible.*

Proof. Suppose that a is prime, and that $a = bc$. We want to prove that either b or c is a unit. By definition of prime, we must have that a divides either b or c . Let us say that a divides b . Thus

$$b = ad \Rightarrow b = bcd \Rightarrow b(1 - cd) = 0 \Rightarrow cd = 1$$

using that R is an integral domain, and thus c is a unit. The same argument works if we assume that a divides c , and we conclude that a is irreducible. \square

Example 2.6. Consider the ring

$$R = \mathbb{Z}[\sqrt{-3}] = \{a + ib\sqrt{3}, a, b \in \mathbb{Z}\}.$$

We want to see that 2 is irreducible but not prime.

- Let us first check that 2 is indeed irreducible. Suppose that

$$2 = (a + ib\sqrt{3})(c + id\sqrt{3}).$$

Since 2 is real, it is equal to its conjugate, and thus

$$2\bar{2} = (a + ib\sqrt{3})(c + id\sqrt{3})(a - ib\sqrt{3})(c - id\sqrt{3})$$

implies that

$$4 = (a^2 + 3b^2)(c^2 + 3d^2).$$

We deduce that $a^2 + 3b^2$ must divide 4, and it cannot possibly be 2, since we have a sum of squares in \mathbb{Z} . If $a^2 + 3b^2 = 4$, then $c^2 + 3d^2 = 1$ and $d = 0$, $c = \pm 1$. Vice versa if $c^2 + 3d^2 = 4$ then $a^2 + 3b^2 = 1$, and $b = 0$, $a = \pm 1$. In both cases we get that one of the factors of 2 is unit, namely ± 1 .

- We now have to see that 2 is not a prime. Clearly

$$2 \mid (1 + i\sqrt{3})(1 - i\sqrt{3}) = 4.$$

But 2 divides neither $1 + i\sqrt{3}$ nor $1 - i\sqrt{3}$.

We can see from the above example that the problem which arises is the lack of unique factorization.

Definition 2.22. A **unique factorization domain (UFD)** is an integral domain R satisfying that

1. every element $0 \neq a \in R$ can be written as a product of irreducible factors p_1, \dots, p_n up to a unit u , namely:

$$a = up_1 \dots p_n.$$

2. The above factorization is unique, that is, if

$$a = up_1 \dots p_n = vq_1 \dots q_m$$

are two factorizations into irreducible factors p_i and q_j with units u, v , then $n = m$ and p_i and q_i are associate for all i .

We now prove that the distinction between irreducible and prime disappear in a unique factorization domain.

Proposition 2.14. *In a unique factorization domain R , we have that a is irreducible if and only if a is prime.*

Proof. We already know that prime implies irreducible. Let us show that now, we also have irreducible implies prime.

Take a to be irreducible and assume that $a \mid bc$. This means that $bc = ad$ for some $d \in R$. Using the property of unique factorization, we decompose d, b and c into products of irreducible terms (resp. d_i, b_i, c_i up to units u, v, w):

$$a \cdot ud_1 \dots d_r = vb_1 \dots b_s \cdot wc_1 \dots c_t.$$

Since the factorization is unique, a must be associate to some either b_i or c_i , implying that a divides b or c , which concludes the proof. \square

We now want to connect the property of unique factorization to ideals.



Figure 2.1: Amalie Emmy Noether (1882-1935)

Definition 2.23. Let a_1, a_2, \dots be elements of an integral domain R . If the sequence of principal ideals

$$(a_1) \subseteq (a_2) \subseteq (a_3) \subseteq \dots$$

stabilizes, i.e., we have

$$(a_n) = (a_{n+1}) = \dots$$

for some n , then we say that R satisfies the [ascending chain condition on principal ideals](#).

If the same condition holds but for general ideals, not necessarily principal, we call R a [Noetherian ring](#), in honor of the mathematician Emmy Noether.

Theorem 2.15. *Let R be an integral domain.*

1. *If R is a UFD, then R satisfies the ascending chain condition on principal ideals.*
2. *If R satisfies the ascending chain condition on principal ideals, then every non-zero element of R can be factored into irreducible (this says nothing about the unicity of the factorization).*
3. *If R is such that every non-zero element of R can be factored into irreducible, and in addition every irreducible element is prime, then R is a UFD.*

Thus R is a UFD if and only if it satisfies the ascending chain condition on principal ideals and every irreducible element of R is prime.

Proof. 1. Recall that in a UFD, prime and irreducible are equivalent. Consider an ascending chain of principal ideals

$$(a_1) \subseteq (a_2) \subseteq (a_3) \subseteq \dots$$

We have that $a_{i+1} \mid a_i$ for all i . Thus the prime factors of a_{i+1} consist of some (possibly all) prime factors of a_i . Since a_1 has a unique factorization into finitely many prime factors, the prime factors will end up being the same, and the chain will stabilize.

2. Take $0 \neq a_1 \in R$. If a_1 is irreducible, we are done. Let us thus assume that a_1 is not irreducible, that is

$$a_1 = a_2 b_2$$

where a_2 and b_2 are not unit. Since $a_2 \mid a_1$, we have $(a_1) \subseteq (a_2)$, and actually

$$(a_1) \subsetneq (a_2).$$

Indeed, if $(a_1) = (a_2)$, then a_2 would be a multiple of a_1 , namely $a_2 = ca_1$ and thus

$$a_1 = a_2 b_2 \Rightarrow a_1 = ca_1 b_2 \Rightarrow a_1(1 - cb_2) = 0$$

implying that $cb_2 = 1$ and thus b_2 is a unit. This contradicts the assumption that a_1 is not irreducible. This computation has shown us that whenever we get a factor which is not irreducible, we can add a new principal ideal to the chain of ideals. Thus, if $a_2 b_2$ is a product of irreducible, we are done. Otherwise, we have that say a_2 is not irreducible, and $a_2 = a_3 b_3$, yielding

$$(a_1) \subsetneq (a_2) \subsetneq (a_3).$$

Since R satisfies the ascending chain condition on principal ideals, this process cannot go on and must stop, showing that we have a factorization into irreducible.

3. We now know that R allows a factorization into irreducible. We want to prove that this factorization is unique, under the assumption that every irreducible is prime. Suppose thus that

$$a = up_1 p_2 \cdots p_n = vq_1 q_2 \cdots q_m$$

where u, v are units and p_i, q_j are irreducible. p_1 is an irreducible but also a prime by assumption, thus it must divide one of the q_j , say q_1 , and we have $q_1 = p_1 d$. Since q_1 is irreducible, d must be a unit, and q_1 and p_1 are associate. We can iterate the process to find that q_i and p_i are associate for all i .

□

We now introduce a notion stronger than being a unique factorization domain.

Definition 2.24. A **principal ideal domain** (PID) is an integral domain in which every ideal is principal.

Theorem 2.16. A principal ideal domain R is a unique factorization domain.

Proof. What we will prove is that if R is a principal ideal domain, then

- R satisfies the ascending chain condition on principal ideals.
- every irreducible in R is also prime.

Having proved these two claims, we can conclude using the above theorem.

Let us first prove that R satisfies the ascending chain condition on principal ideals. Consider the following sequence of principal ideals

$$(a_1) \subseteq (a_2) \subseteq (a_3) \dots$$

and let $\mathcal{I} = \cup_{i=1}^{\infty} (a_i)$. Note that \mathcal{I} is an ideal of R (be careful, a union of ideals is not an ideal in general!). Indeed, we have that \mathcal{I} is closed under addition: take $a, b \in \mathcal{I}$, then there are ideals (a_j) and (a_k) in the chain with $a \in (a_j)$ and $b \in (a_k)$. If $m \geq \max(j, k)$, then both $a, b \in (a_m)$ and so do $a+b$. To check that \mathcal{I} is closed under multiplication by an element of R , take again $a \in \mathcal{I}$. Then $a \in (a_j)$ for some j . If $r \in R$, then $ra \in (a_j)$ implying that $ra \in \mathcal{I}$.

Now by assumption, \mathcal{I} is a principal ideal, generated by, say b : $\mathcal{I} = (b)$. Since b belongs to $\cup_{i=1}^{\infty} (a_i)$, it must belong to some (a_n) . Thus $\mathcal{I} = (b) \subseteq (a_n)$. For $j \geq n$, we have

$$(a_j) \subseteq \mathcal{I} \subseteq (a_n) \subseteq (a_j)$$

which proves that the chain of ideal stabilizes.

We are left to prove that every irreducible element is also prime. Let thus a be an irreducible element. Consider the principal ideal (a) generated by a . Note that (a) is a proper ideal: if $(a) = R$, then $1 \in (a)$ and thus a is a unit, which is a contradiction.

We have that (a) is included in a maximal ideal \mathcal{I} (this can be deduced from either the ascending chain condition or from the theorem (Krull's theorem) that proves that every ideal is contained in a maximal ideal). Since R is a principal ideal domain, we have that $\mathcal{I} = (b)$. Thus

$$(a) \subseteq (b) \Rightarrow b \mid a \Rightarrow a = bd$$

where a is irreducible, b cannot be a unit (since \mathcal{I} is by definition of maximal ideal a proper ideal), and thus d has to be a unit of R . In other words, a and b are associate. Thus

$$(a) = \mathcal{I} = (b).$$

Since \mathcal{I} is a maximal ideal, it is prime implying that a is prime, which concludes the proof. \square

Determining whether a ring is a principal ideal domain is in general quite a tough question. It is still an open conjecture (called [Gauss's conjecture](#)) to decide whether there are infinitely many real quadratic fields which are principal (we use the terminology “principal” for quadratic fields by abuse of notation, it actually refers to their ring of integers, that is rings of the form either $\mathbb{Z}[\sqrt{d}]$ if $d \equiv 2$ or $3 \pmod{4}$ or $\mathbb{Z}[\frac{1+\sqrt{d}}{2}]$ else).

One way mathematicians have found to approach this question is to actually prove a stronger property, namely whether a ring R is Euclidean.

Definition 2.25. Let R be an integral domain. We say that R is a [Euclidean domain](#) if there is a function Ψ from $R \setminus \{0\}$ to the non-negative integers such that

$$a = bq + r, \quad a, b \in R, \quad b \neq 0, \quad q, r \in R$$

where either $r = 0$ or $\Psi(r) < \Psi(b)$.

When the division is performed with natural numbers, it is clear what it means that $r < b$. When we work with polynomials instead, we can say that $\deg r < \deg b$. The function Ψ generalizes these notions.

Theorem 2.17. *If R is a Euclidean domain, then R is a principal ideal domain.*

Proof. Let \mathcal{I} be an ideal of R . If $\mathcal{I} = \{0\}$, it is principal and we are done. Let us thus take $\mathcal{I} \neq \{0\}$. Consider the set

$$\{\Psi(b), b \in \mathcal{I}, b \neq 0\}.$$

It is included in the non-negative integers by definition of Ψ , thus it contains a smallest element, say n . Let $0 \neq b \in \mathcal{I}$ such that $\Psi(b) = n$.

We will now prove that $\mathcal{I} = (b)$. Indeed, take $a \in \mathcal{I}$, and compute

$$a = bq + r$$

where $r = 0$ or $\Psi(r) < \Psi(b)$. This yields

$$r = a - bq \in \mathcal{I}$$

and $\Psi(r) < \Psi(b)$ cannot possibly happen by minimality of n , forcing r to be zero. This concludes the proof. \square

Example 2.7. Consider the ring

$$\mathbb{Z}[\sqrt{d}] = \{a + b\sqrt{d}, a, b \in \mathbb{Z}\}$$

with

$$\Psi(a + b\sqrt{d}) = |a^2 - b^2d|.$$

We will show that we have a Euclidean domain for $d = -2, -1, 2$.

Note that $\mathbb{Z}[\sqrt{d}]$ is an integral domain. Take $\alpha, \beta \neq 0$ in $\mathbb{Z}[\sqrt{d}]$. Now we would like to perform the division of α by β to get something of the form

$$\alpha = \beta q + r, \quad q, r \in \mathbb{Z}[\sqrt{d}].$$

Since $\mathbb{Z}[\sqrt{d}]$ is not a field, there is no reason for this division to give a result in $\mathbb{Z}[\sqrt{d}]$ (that is, $q, r \in \mathbb{Z}[\sqrt{d}]$), however, we can compute the division in $\mathbb{Q}(\sqrt{d})$:

$$\alpha/\beta = q',$$

with $q' = x + \sqrt{d}y$ with x, y rational. Let us now approximate x, y by integers x_0, y_0 , namely take x_0, y_0 such that

$$|x - x_0| \leq 1/2, |y - y_0| \leq 1/2.$$

Take

$$q = x_0 + y_0\sqrt{d}, r = \beta((x - x_0) + (y - y_0)\sqrt{d}),$$

where clearly $q \in \mathbb{Z}[\sqrt{d}]$, then

$$\begin{aligned} \beta q + r &= \beta(x_0 + y_0\sqrt{d}) + \beta((x - x_0) + (y - y_0)\sqrt{d}) \\ &= \beta(x + y\sqrt{d}) = \beta q' = \alpha, \end{aligned}$$

which at the same time shows that $r \in \mathbb{Z}[\sqrt{d}]$. We are left to show that $\Psi(r) < \Psi(\beta)$. We have

$$\begin{aligned} \Psi(r) &= \Psi(\beta)\Psi((x - x_0) + (y - y_0)\sqrt{d}) \\ &= \Psi(\beta)|(x - x_0)^2 - d(y - y_0)^2| \\ &\leq \Psi(\beta)[|x - x_0|^2 + |d||y - y_0|^2] \\ &\leq \Psi(\beta) \left(\frac{1}{4} + |d|\frac{1}{4} \right) \end{aligned}$$

showing that $\mathbb{Z}[\sqrt{d}]$ is indeed a Euclidean domain for $d = -2, -1, 2$.

Below is a summary of the ring hierarchy (recall that PID and UFD stand respectively for principal ideal domain and unique factorization domain):

integral domains \supset UFD \supset PID \supset Euclidean domains
--

Note that though the Euclidean division may sound like an elementary concept, as soon as the ring we consider is fancier than \mathbb{Z} , it becomes quickly a difficult problem. We can see that from the fact that being Euclidean is stronger than being a principal ideal domain. All the inclusions are strict, since one may check that $\mathbb{Z}[\sqrt{-3}]$ is an integral domain but is not a UFD, $\mathbb{Z}[X]$ is a UFD which is not PID, while $\mathbb{Z}[(1 + i\sqrt{19})/2]$ is a PID which is not a Euclidean domain.

2.7 Irreducible polynomials

Recall the definition of irreducible that we have seen: a non-zero element a which is not a unit is said to be irreducible if $a = bc$ implies that either b or c is a unit. Let us focus on the case where the ring is a ring of polynomials $R[X]$ and R is an integral domain.

Definition 2.26. If R is an integral domain, then an irreducible element of $R[X]$ is called an **irreducible polynomial**.

In the case of a field F , then units of $F[X]$ are non-zero elements of F . Then we get the more familiar definition that an irreducible element of $F[X]$ is a polynomial of degree at least 1, that cannot be factored into two polynomials of lower degree.

Let us now consider the more general case where R is an integral domain (thus not necessarily a field, it may not even be a unique factorization domain). To study when polynomials over an integral domain R are irreducible, it is often more convenient to place oneself in a suitable field that contains R , since division in R can be problematic. To do so, we will now introduce the field of fractions, also called quotient field, of R . Since there is not much more difficulty in treating the general case, that is, when R is a commutative ring, we present this construction.

Let S be a subset of R which is closed under multiplication, contains 1 and does not contain 0. This definition includes the set of all non-zero elements of an integral domain, or the set of all non-zero elements of a commutative ring that are not zero divisors. We define the following equivalence relation on $R \times S$:

$$(a, b) \sim (c, d) \iff s(ad - bc) = 0 \text{ for some } s \in S.$$

It is clearly reflexive and symmetric. Let us check the transitivity. Suppose that $(a, b) \sim (c, d)$ and $(c, d) \sim (e, f)$. Then

$$s(ad - bc) = 0 \text{ and } t(cf - de) = 0$$

for some $s, t \in S$. We can now multiply the first equation by tf , the second by sb and add them

$$stf(ad - bc) + tbs(cf - de) = 0$$

to get

$$sdt(fa - be) = 0$$

which proves the transitivity.

What we are trying to do here is to mimic the way we deal with \mathbb{Z} . If we take non-zero $a, b, c, d \in \mathbb{Z}$, we can write down $a/b = c/d$, or equivalently $ad = bc$, which is also what $(a, b) \sim (c, d)$ satisfies by definition if we take R to be an integral domain. In a sense, (a, b) is some approximation of a/b .

Formally, if $a \in R$ and $b \in S$, we define the fraction a/b to be the equivalence class of the pair (a, b) . The set of all equivalence classes is denoted by $S^{-1}R$. To make it into a ring, we define the following laws in a natural way:

- addition:

$$\frac{a}{b} + \frac{c}{d} = \frac{ad + bc}{bd}.$$

- multiplication:

$$\frac{a}{b} \frac{c}{d} = \frac{ac}{bd}.$$

- additive identity:

$$\frac{0}{1} = \frac{0}{s}, \quad s \in S.$$

- additive inverse:

$$-\frac{a}{b} = \frac{-a}{b}.$$

- multiplicative identity:

$$\frac{1}{1} = \frac{s}{s}, \quad s \in S.$$

To prove that we really obtain a ring, we need to check that all these laws are well-defined.

Theorem 2.18. *With the above definitions, the set of equivalence classes $S^{-1}R$ is a commutative ring.*

1. *If R is an integral domain, so is $S^{-1}R$.*
2. *If R is an integral domain, and $S = R \setminus \{0\}$, then $S^{-1}R$ is a field.*

Proof. Addition is well-defined. If $a_1/b_1 = c_1/d_1$ and $a_2/b_2 = c_2/d_2$, then for some $s, t \in S$, we have

$$s(a_1d_1 - b_1c_1) = 0 \text{ and } t(a_2d_2 - b_2c_2) = 0.$$

We can now multiply the first equation by tb_2d_2 and the second by sb_1d_1 to get

$$tb_2d_2s(a_1d_1 - b_1c_1) = 0 \text{ and } sb_1d_1t(a_2d_2 - b_2c_2) = 0,$$

and adding them yields

$$st[d_2d_1(b_2a_1 + b_1a_2) - b_2b_1(d_2c_1 + d_1c_2)] = 0$$

that is

$$\frac{b_2a_1 + b_1a_2}{b_2b_1} = \frac{d_2c_1 + d_1c_2}{d_2d_1},$$

which can be rewritten as

$$\frac{a_1}{b_1} + \frac{a_2}{b_2} = \frac{c_1}{d_1} + \frac{c_2}{d_2}$$

and we conclude that addition does not depend on the choice of a representative in an equivalence class.

Multiplication is well-defined. We start as before. If $a_1/b_1 = c_1/d_1$ and $a_2/b_2 = c_2/d_2$, then for some $s, t \in S$, we have

$$s(a_1d_1 - b_1c_1) = 0 \text{ and } t(a_2d_2 - b_2c_2) = 0.$$

Now we multiply instead the first equation by ta_2d_2 , the second by sc_1b_1 and we add them:

$$st[a_2d_2a_1d_1 - c_1b_1b_2c_2] = 0.$$

This implies, as desired, that

$$\frac{a_1a_2}{b_1b_2} = \frac{c_1c_2}{d_1d_2}.$$

To be complete, one should check that the properties of a ring are fulfilled, but this follows from the fact that addition and multiplication are carried the usual way.

1. We want to prove that $S^{-1}R$ is an integral domain. We assume that R is an integral domain, and we need to check the definition of an integral domain for $S^{-1}R$. Namely, suppose that $(a/b)(c/d) = 0$ in $S^{-1}R$, that is

$$\frac{ac}{bd} = \frac{0}{1}.$$

This means that $(ac, bd) \sim (0, 1)$ and $acs = 0$ for some $s \in S$. Now $acs = 0$ is an equation in R , which is an integral domain, and $s \neq 0$, thus $ac = 0$, so either a or c is 0, and consequently either a/b or c/d is zero.

2. To conclude, we want to prove that $S^{-1}R$ is a field, assuming that R is an integral domain, and $S = R \setminus \{0\}$. We consider a/b a non-zero element of $S^{-1}R$, for which we need to find an inverse. Note that a and b are non-zero, thus they are both in S meaning that both a/b and b/a are in $S^{-1}R$ and b/a is the multiplicative inverse of a/b .

□

Definition 2.27. Let R be a commutative ring. Based on the above, the set of equivalence classes $S^{-1}R$ is a commutative ring, called the **ring of fractions** of R by S . If R is an integral domain, and $S = R \setminus \{0\}$, then $S^{-1}R$ is called the **field of fractions** or **quotient field** of R .

Now that we have defined a suitable field, we are left to prove that we can embed an integral domain R in its quotient field.

Proposition 2.19. *A commutative ring R can be embedded in its ring of fractions $S^{-1}R$, where S is the set of all its non-divisors of zero. In particular, an integral domain can be embedded in its quotient field, which is furthermore the smallest field containing R .*

Proof. Consider the following map:

$$f : R \rightarrow S^{-1}R, \quad a \mapsto f(a) = a/1.$$

It is not hard to check that f is a ring homomorphism. If S has no zero divisor, we have that the kernel of f is given by the set of a such that $f(a) = a/1 = 0/1$, that is the set of a such that $sa = 0$ for some s . Since s is not a zero divisor, we have $a = 0$ and f is a monomorphism. \square

Let us get back to the irreducible polynomials, and consider now the case where D is a unique factorization domain. It is not necessarily a field, but we now know how to embed it in a suitable field, namely its field of fractions, or quotient field. Take the polynomial $f(X) = a + abX$, $a \neq 0$ not a unit. Since we can factor it as

$$f(X) = a(1 + bX)$$

where a is not a unit by assumption, this polynomial is not irreducible. But we do not really have a factorization into two polynomials of lower degree. What happens here is that the constant polynomials are not necessarily units, unlike in the case of fields. To distinguish this case, we introduce the notion of primitive polynomial.

Definition 2.28. Let D be a unique factorization domain and let $f \in D[X]$. We call the greatest common divisor of all the coefficients of f the **content** of f , denoted by $c(f)$. A polynomial whose content is a unit is called a **primitive polynomial**.

We can now rule out the above example, and we will prove later that this allows us to say that a primitive polynomial is irreducible if and only if it cannot be factored into two polynomials of lower degree. Be careful however that “primitive polynomial” has a different meaning if it is defined over a field.

The next goal is to prove Gauss lemma, which in particular implies that the product of two primitive polynomials is a primitive polynomial.

We start with a lemma.

Lemma 2.20. Let D be a unique factorization domain, and consider $f \neq 0, g, h \in D[X]$ such that $pf(X) = g(X)h(X)$ with p a prime. Then either p divides all the coefficients of g or p divides all the coefficients of h .

Before starting the proof, let us notice that this lemma is somehow a generalization of the notion of prime. Instead of saying that $p|ab$ implies $p|a$ or $p|b$, we have $p|g(X)h(X)$ implies that $p|g(X)$ or $p|h(X)$ (dividing the whole polynomial means dividing all of its coefficients).

Proof. Denote

$$g(X) = g_0 + g_1X + \dots + g_sX^s, \quad h(X) = h_0 + h_1X + \dots + h_tX^t.$$

Suppose by contradiction that p does not divide all coefficients of g and does not divide all coefficients of h either. Then let g_u and h_v be the coefficients of



Figure 2.2: Carl Friedrich Gauss (1777-1855)

minimum index not divisible by p . Then the coefficient of X^{u+v} in $g(X)h(X)$ is

$$g_0h_{u+v} + g_1h_{u+v-1} + \dots + g_uh_v + \dots + g_{u+v-1}h_1 + g_{u+v}h_0.$$

By definition of u and v , p divides every term but g_uh_v , thus p cannot possibly divide the entire expression, and thus there exists a coefficient of $g(X)h(X)$ not divisible by p . This contradicts the fact that $p|g(X)h(X)$. \square

Proposition 2.21. (Gauss Lemma). *Let f, g be non-constant polynomials in $D[X]$ where D is a unique factorization domain. The content of a product of polynomials is the product of the contents, namely*

$$c(fg) = c(f)c(g),$$

up to associates. In particular, the product of two primitive polynomials is primitive.

Proof. Let us start by noticing that by definition of content, we can rewrite

$$f(X) = c(f)f^*(X), \quad g(X) = c(g)g^*(X),$$

where $f^*, g^* \in D[X]$ are primitive. Clearly

$$fg = c(f)c(g)f^*g^*.$$

Since $c(f)c(g)$ divides fg , it divides every coefficient of fg and thus their greatest common divisor:

$$c(f)c(g) \mid c(gf).$$

We now prove the converse, namely that $c(gf) \mid c(g)c(f)$. To do that, we consider each prime p appearing in the factorization of $c(gf)$ and argue that

$p \mid c(f)c(g)$. Let thus p be a prime factor of $c(fg)$. Since $fg = c(fg)(fg)^*$, we have that $c(fg)$ divides fg , that is

$$p \mid c(f)c(g)f^*g^* = (c(f)f^*)(c(g)g^*).$$

By the above lemma, either $p \mid c(f)f^*$ or $p \mid c(g)g^*$, say $p \mid c(f)f^*$, meaning that either $p \mid c(f)$ or $p \mid f^*$. Since f^* is primitive, p cannot possibly divide f^* , and thus

$$p \mid c(f) \Rightarrow p \mid c(f)c(g).$$

If p appears with multiplicity, we iterate the reasoning with the same p . \square

We are now ready to connect irreducibility over a unique factorization domain and irreducibility over the corresponding quotient field or field of fractions.

Proposition 2.22. *Let D be a unique factorization domain with quotient field F . If f is a non-constant polynomial in $D[X]$, then f is irreducible over D if and only if f is primitive and irreducible over F .*

Proof. First assume that f is irreducible over D .

f is primitive. Indeed, if f were not primitive, then we could write

$$f = c(f)f^*,$$

where $c(f)$ denotes the content of f and f^* is primitive. Since we assume f is not primitive, its content cannot be a unit, which contradicts the irreducibility of f over D , and we conclude that f is primitive.

f is irreducible over F . Again assume by contradiction that f is not irreducible over F . Now F is a field, thus reducible means f can be factored into a product of two non-constant polynomials in $F[X]$ of smaller degree:

$$f(X) = g(X)h(X), \quad \deg g < \deg f, \quad \deg h < \deg f.$$

Since g, h are in $F[X]$, and F is the field of fractions of D , we can write

$$g(X) = \frac{a}{b}g^*(X), \quad h(X) = \frac{c}{d}h^*(X), \quad a, b, c, d \in D$$

and g^*, h^* primitive. Thus

$$f(X) = \frac{ac}{bd}g^*(X)h^*(X)$$

where g^*h^* is a primitive polynomial by Gauss Lemma. Since we have already proven (in the 1st part) that f is primitive, it must be that $ac/bd = u$ is a unit. But this would mean that

$$f(X) = ug^*(X)h^*(X)$$

which contradicts the fact that $f(X)$ is irreducible over $D[X]$ and we conclude that f is also irreducible over $F[X]$.



Figure 2.3: Ferdinand Eisenstein (1823-1852)

We are left to prove the converse. Let then f be a primitive and irreducible polynomial over F . We do it by contraction, and assume that the primitive polynomial f is not irreducible over D :

$$f(X) = g(X)h(X).$$

Since f is primitive, $\deg g$ and $\deg h$ are at least 1. But then neither g nor h can be a unit in $F[X]$ (these are units in F) and thus

$$f = gh$$

contradicts the irreducibility of f over F . □

In other words, we have proven that f irreducible over D is equivalent to f primitive and cannot be factored into two polynomials of lower degree in $F[X]$.

To conclude, we present a practical criterion to decide whether a polynomial in $D[X]$ is irreducible over F .

Proposition 2.23. (Eisenstein's criterion). *Let D be a unique factorization domain, with quotient field F and let*

$$f(X) = a_n X^n + \dots + a_1 X + a_0$$

be a polynomial in $D[X]$ with $n \geq 1$ and $a_n \neq 0$.

If p is a prime in D and p divides a_i , $0 \leq i < n$ but p does not divide a_n nor does p^2 divide a_0 , then f is irreducible over F .

Proof. We first divide f by its content, to get a primitive polynomial. By the above proposition, it is enough to prove that this primitive polynomial is irreducible over D .

Let thus f be a primitive polynomial and assume by contradiction it is reducible, that is

$$f(X) = g(X)h(X)$$

with

$$g(X) = g_0 + \dots + g_r X^r, \quad h(X) = h_0 + \dots + h_s X^s.$$

Notice that r cannot be zero, for if $r = 0$, then $g_0 = g$ would divide f and thus all a_i implying that g_0 divides the content of f and is thus a unit. But this would contradict the fact that f is reducible. We may from now on assume that

$$r \geq 1, \quad s \geq 1.$$

Now by hypothesis, $p \mid a_0 = g_0 h_0$ but p^2 does not divide a_0 , meaning that p cannot divide both g_0 and h_0 . Let us say that

$$p \mid g_0$$

and p does not divide h_0 (and vice-versa).

By looking at the dominant coefficient $a_n = g_r h_s$, we deduce from the assumption that p does not divide a_n that p cannot possibly divide g_r . Let i be the smallest integer such that p does not divide g_i . Then

$$1 \leq i \leq r < n = r + s.$$

Let us look at the i th coefficient

$$a_i = g_0 h_i + g_1 h_{i-1} + \dots + g_i h_0$$

and by choice of i , p must divide g_0, \dots, g_{i-1} . Since p divides a_i by assumption, it thus must divide the last term $g_i h_0$, and either $p \mid g_i$ or $p \mid h_0$ by definition of prime. Both are impossible: we have chosen p dividing neither h_0 nor g_i . This concludes the proof. \square

The main definitions and results of this chapter are

- **(2.1-2.2).** Definitions of: ring, zero divisor, unit, integral domain, division ring, subring, characteristic, ring homomorphism, ideal, quotient ring. Factor and 1st Isomorphism Theorem for rings.
- **(2.3-2.4).** Operations on ideals, Chinese Remainder Theorem, Correspondence Theorem for rings. Definitions of: principal ideal, maximal ideal, prime ideal, the characterization of the two latter in the commutative case.
- **(2.5).** Polynomial Euclidean division, number of roots of a polynomial.
- **(2.6).** Definitions of: associate, prime, irreducible, unique factorization domain, ascending chain condition, principal ideal domain, Euclidean domain. Connections between prime and irreducible. Hierarchy among UFD, PID and Euclidean domains.
- **(2.7).** Construction of ring of fractions. Definitions of: content of a polynomial, primitive polynomial. Gauss Lemma, Eisenstein's criterion.

Chapter 3

Field Theory

Abstract field theory emerged from three theories, which we would now call Galois theory, algebraic number theory and algebraic geometry.

Field theoretic notions appeared, even though still implicitly, in the modern theory of solvability of polynomial equations, as introduced by Abel and Galois in the early nineteenth century. Galois had a good insight into fields obtained by adjoining roots of polynomials, and he proved what we call now the Primitive Element Theorem.

Independently, Dedekind and Kronecker came up with the notion of algebraic number fields, arising from three major number -theoretic problems: Fermat's Last Theorem, reciprocity laws and representation of integers by binary quadratic forms.

Algebraic geometry is the study of algebraic curves and their generalizations to higher dimensions, namely, algebraic varieties. Dedekind and Weber carried over to algebraic functions the ideas which Dedekind had earlier introduced for algebraic numbers, that is, define an algebraic function field as a finite extension of the field of rational functions.

At the end of the nineteenth century, abstraction and axiomatics started to take place. Cantor (1883) defined the real numbers as equivalence classes of Cauchy sequences, von Dyck (1882) gave an abstract definition of group (about thirty years after Cayley had defined a finite group). Weber's definition of a field appeared in 1893, for which he gave number fields and function fields as examples. In 1899, Hensel initiated a study of p -adic numbers, taking as starting point the analogy between function fields and number fields. It is the work of Steinitz in 1910 that initiated the abstract study of fields as an independent subject. A few examples of his results are: classification of fields into those of characteristic zero and those of characteristic p , development of the theory of transcendental extensions, recognition that it is precisely the finite, normal, separable extensions to which Galois theory applies, proof of the existence of the algebraic closure of any field.

Major developments in field theory and related areas that followed Steinitz's work include valuation theory, class field theory, infinite Galois theory and finite fields.

3.1 Field extension and minimal polynomial

Definition 3.1. If F and E are fields, and $F \subseteq E$, we say that E is an **extension** of F , and we write either $F \leq E$ or E/F .

Examples 3.1. Here are some classical examples:

1. $\mathbb{C} = \{a + bi, a, b \in \mathbb{R}\}$ is a field extension of \mathbb{R} .
2. $\mathbb{Q}(\sqrt{2}) = \{a + b\sqrt{2}, a, b \in \mathbb{Q}\}$ is a field extension of \mathbb{Q} .
3. $\mathbb{Q}(i) = \{a + bi, a, b \in \mathbb{Q}\}$ is a field extension of \mathbb{Q} .

If E is an extension of F , then in particular E is an abelian group under addition, and we may multiply $x \in E$ by $\lambda \in F$. We can see that this endows E with a structure of F -vector space (the elements of E are seen as vectors, those of F as scalars). It then makes sense to speak of the dimension of E over F .

Definition 3.2. Let E/F be a field extension. The dimension of E as F -vector space is called the **degree** of the extension, written $[E : F]$. If $[E : F] < \infty$, we say that E is a finite extension of F , or that the extension E/F is finite.

Let us get back to our examples:

- Examples 3.2.**
1. Consider the field extension \mathbb{C}/\mathbb{R} . We have that \mathbb{C} is a vector space of dimension 2 over \mathbb{R} . It is thus an extension of degree 2 (with basis $\{1, i\}$).
 2. The field extension $\mathbb{Q}(\sqrt{2})/\mathbb{Q}$ is of degree 2, it is called a **quadratic extension** of \mathbb{Q} .
 3. The field extension $\mathbb{Q}(i)/\mathbb{Q}$ is also a quadratic field extension of \mathbb{Q} .
 4. Both $\mathbb{Q}(\sqrt{2})/\mathbb{Q}$ and $\mathbb{Q}(i)/\mathbb{Q}$ are finite field extensions of \mathbb{Q} . Finite extensions of \mathbb{Q} are called **number fields**.

If we look at \mathbb{C} , we see it is obtained by adding i to \mathbb{R} , and i is a root of the polynomial $X^2 + 1$. Similarly, $\mathbb{Q}(\sqrt{2})/\mathbb{Q}$ is obtained by adding a root of the polynomial $X^2 - 2$. In what follows, we will make formal the connection between roots of polynomials and field extensions.

Before we start, recall that if we have two fields E, F and a **field homomorphism** between them (that is, a ring homomorphism between two fields), then f is a monomorphism. We have seen the argument in the previous chapter already: the kernel of a ring homomorphism is an ideal, and a field has only trivial ideals, namely $\{0\}$ and itself, and it cannot be that the whole field is the kernel.

Theorem 3.1. *Let f be a non-constant polynomial over a field F . Then there is an extension E/F and an element $\alpha \in E$ such that $f(\alpha) = 0$.*

Proof. Recall that $F[X]$ is a unique factorization domain, thus f can be factored into a product of irreducible polynomials, and we may assume without loss of generality that f is itself irreducible. Consider now the ideal

$$\mathcal{I} = (f(X))$$

in $F[X]$, the ring of polynomials with indeterminate X and coefficients in F . Again using that $F[X]$ is a unique factorization domain, we have that $f(X)$ is irreducible and equivalently prime, implying that $(f(X))$ is prime. Now $F[X]$ is furthermore a principal ideal domain. This means that $\mathcal{I} = (f(X))$ is contained in a maximal ideal $(q(X))$, so that $q(X)$ divides the prime $f(X)$. Since $q(X)$ cannot be a unit, $f(X)$ and $q(X)$ are associates, and $(f(X)) = (q(X))$, proving that $(p(X)) = \mathcal{I}$ is maximal. Thus by the characterization of maximal ideals with respect to their quotient ring, we have that

$$E = F[X]/\mathcal{I}$$

is a field. We now place an isomorphic copy of F inside E via the monomorphism

$$h : F \rightarrow E, \quad a \mapsto a + \mathcal{I}.$$

This thus gives a field extension E/F . Now let

$$\alpha = X + \mathcal{I} \in E.$$

We are left to prove that α is a root of $f(X)$. If $f(X) = a_0 + a_1X + \dots + a_nX^n$, then

$$\begin{aligned} f(\alpha) &= (a_0 + \mathcal{I}) + a_1(X + \mathcal{I}) + \dots + a_n(X + \mathcal{I})^n \\ &= a_0 + \mathcal{I} + a_1X + a_1\mathcal{I} + \dots + a_nX^n + \dots + a_n\mathcal{I}^n \\ &= (a_0 + a_1X + \dots + a_nX^n) + \mathcal{I} \\ &= f(X) + \mathcal{I} \end{aligned}$$

which is zero in E . □

The extension E is sometimes said to be obtained from F by adjoining a root of f .

Remark. Note that in the above proof, we have shown that a prime ideal in a principal ideal domain is maximal. This allows to prove in general that a ring is a principal ideal domain if and only if it is a unique factorization domain and every prime ideal is maximal.

Definition 3.3. If E is an extension of F , an element $\alpha \in E$ is said to be **algebraic** over F if there is a non-constant polynomial $f \in F[X]$ such that $f(\alpha) = 0$. If α is not algebraic over F , it is said to be **transcendental** over F . If every element of E is algebraic over F , then E is said to be an **algebraic extension** of F .

Suppose that $\alpha \in E$ is algebraic over F . Thus there exists by definition a polynomial $f \in F[X]$ with $f(\alpha) = 0$. It thus makes sense to consider the set \mathcal{I} of all polynomials $g \in F[X]$ such that $g(\alpha) = 0$. Clearly

- if g_1, g_2 are in \mathcal{I} , so does $g_1 \pm g_2$,
- if $g \in \mathcal{I}$ and $h \in F[X]$, then $gh \in \mathcal{I}$.

This tells us that $\mathcal{I} = \{g \in F[X], g(\alpha) = 0\}$ is an ideal of $F[X]$.

Since $F[X]$ is a principal ideal domain, we have

$$\mathcal{I} = (m(X))$$

for some $m(X)$ in $F[X]$. Any two generators of \mathcal{I} are thus multiple of each others, so if $m(X)$ is monic, it has to be unique. This polynomial $m(X)$ has the following properties:

1. If $g \in F[X]$, then $g(\alpha) = 0$ if and only if $m(X)$ divides $g(X)$. This is clear from the definition of \mathcal{I} .
2. $m(X)$ is the monic polynomial of least degree such that $m(\alpha) = 0$, which follows from the above property.
3. $m(X)$ is the unique monic irreducible polynomial such that $m(\alpha) = 0$. Indeed, if $m(X) = h(X)k(X)$ with $\deg h < \deg m$, $\deg k < \deg m$, then either $h(\alpha) = 0$ or $k(\alpha) = 0$, so that either $h(X)$ or $k(X)$ is a multiple of $m(X)$ by the first property, which is impossible. Thus $m(X)$ is irreducible. We are left to prove the unicity of $m(X)$. This comes from the fact that since $m(X)$ is monic, then if there were two irreducible polynomials $m(X)$ and $m'(X)$ such that $m(\alpha) = m'(\alpha) = 0$, they have α as common root, and thus $m(X)$ and $m'(X)$ cannot be distinct. This is the converse of the proposition we will prove below.

Definition 3.4. The polynomial $m(X)$ is called the **minimal polynomial** of α over F . It may be denoted $\min(\alpha, F)$ or $\mu_{\alpha, F}$.

Example 3.3. The polynomial $X^2 + 1$ is the minimal polynomial of i over \mathbb{Q} . It also the minimal polynomial of i over \mathbb{R} .

Proposition 3.2. 1. Let f and g be polynomials over the field F . Then f and g are relatively prime if and only if f and g have no common root in any extension of F .

2. If f and g are distinct monic irreducible polynomials over F , then f and g have no common roots in any extension of F .

Proof. 1. If f and g are relatively prime, their greatest common divisor is 1, so there are polynomials $a(X)$ and $b(X)$ over F such that

$$a(X)f(X) + b(X)g(X) = 1.$$

If there is a common root say α , then we get that $0 = 1$, a contradiction.

Conversely, let us assume that the greatest common divisor $d(X)$ of $f(X)$ and $g(X)$ is non-constant and show that then $f(X)$ and $g(X)$ have a common root. By the above proposition, there exists E an extension of F in which $d(X)$ has a root α . Since $d(X)$ divides both $f(X)$ and $g(X)$, α is a common root of f and g in E .

2. By the first part, it is enough to show that f and g are relatively prime. Assume to the contrary that h is a non-constant divisor of the polynomials f and g which are irreducible. Then $f = f'h$ and $g = g'h$ with f', g' non-zero constant, and $h = \frac{f}{f'} = \frac{g}{g'}$. This is impossible because f and g are monic and distinct.

□

If E is an extension of F and $\alpha \in E$ is a root of a polynomial $f \in F[X]$, one may consider the field $F(\alpha)$ generated by F and α , which is the smallest subfield of E containing both F and α . Alternatively, $F(\alpha)$ can be described as the intersection of all subfields of E containing F and α , or the set of all rational functions

$$\frac{a_0 + a_1\alpha + \cdots + a_m\alpha^m}{b_0 + b_1\alpha + \cdots + b_n\alpha^n}$$

with $a_i, b_j \in F$, $m, n = 0, 1, \dots$ and the denominator is different from 0.

Theorem 3.3. *Let $\alpha \in E$ be algebraic over F , with minimal polynomial $m(X)$ over F of degree n .*

1. We have $F(\alpha) = F[\alpha] = F_{n-1}[\alpha]$ where $F_{n-1}[\alpha]$ denotes the set of all polynomials of degree at most $n-1$ with coefficients in F .
2. $1, \alpha, \dots, \alpha^{n-1}$ form a basis for the vector space $F(\alpha)$ over the field F . Consequently $[F(\alpha) : F] = n$.

Proof. Let us first prove that $F_{n-1}[\alpha]$ is a field. Let $f(X)$ be any non-zero polynomial over F of degree at most $n-1$. Since $m(X)$ is irreducible with $\deg f < \deg m$, $f(X)$ and $m(X)$ are relatively prime, and there exist polynomials $a(X)$ and $b(X)$ over F such

$$a(X)f(X) + b(X)m(X) = 1.$$

Using that α is a root of m , we get

$$a(\alpha)f(\alpha) = 1$$

so that any non-zero element of $F_{n-1}[\alpha]$ has an inverse, and $F_{n-1}[\alpha]$ is a field.

1. Any field containing F and α must contain all polynomials in α , and in particular all those of degree at most $n-1$. Thus

$$F_{n-1}[\alpha] \subset F[\alpha] \subset F(\alpha).$$

But $F(\alpha)$ is the smallest field containing F and α , so

$$F(\alpha) \subset F_{n-1}[\alpha]$$

and we conclude that

$$F(\alpha) = F[\alpha] = F_{n-1}[\alpha].$$

- Now $1, \alpha, \dots, \alpha^{n-1}$ certainly span $F_{n-1}[\alpha]$, and they are linearly independent because if a non-trivial linear combination of them were zero, this would yield a non-zero polynomial of degree less than that of $m(X)$ with α as a root, a contradiction.

□

Example 3.4. Let ζ_5 denote a primitive 5th root of unity (that is, $\zeta_5^5 = 1$ and $\zeta_5^k \neq 1$ for $1 \leq k \leq 4$). We have that $\zeta_5 \in \mathbb{Q}(\zeta_5)$ is algebraic over \mathbb{Q} , with minimal polynomial $X^4 + X^3 + X^2 + X + 1 = 0$ of degree 4 over \mathbb{Q} . A \mathbb{Q} -basis is given by $\{1, \zeta_5, \zeta_5^2, \zeta_5^3\}$ and $[\mathbb{Q}(\zeta_5) : \mathbb{Q}] = 4$.

Once we have a field extension K/F , we can take again K as base field and get another field extension E/K , yielding a tower of extensions $E/K/F$.

Proposition 3.4. Consider the field extensions $E/K/F$.

- If $\alpha_i, i \in I$, form a basis for E over K , and $\beta_j, j \in J$ form a basis for K over F , then $\alpha_i \beta_j, i \in I, j \in J$, form a basis for E over F .
- The degree is multiplicative, namely

$$[E : F] = [E : K][K : F].$$

In particular, $[E : F]$ is finite if and only if $[E : K]$ and $[K : F]$ are finite.

Proof. 1. Take $\gamma \in E$. Then

$$\begin{aligned} \gamma &= \sum_{i \in I} a_i \alpha_i, \quad a_i \in K \\ &= \sum_{i \in I} \left(\sum_{j \in J} b_{ij} \beta_j \right) \alpha_i, \quad b_{ij} \in F. \end{aligned}$$

Thus $\alpha_i \beta_j$ span E over F . We now check the linear independence.

$$\sum_{i,j} \lambda_{ij} \alpha_i \beta_j = 0 \Rightarrow \sum_i \lambda_{ij} \alpha_i = 0$$

for all j and consequently $\lambda_{ij} = 0$ for all i, j which concludes the proof.

- It is enough to use the first part, with

$$[E : K] = |I|, \quad [K : F] = |J|, \quad [E : F] = |I||J|.$$

□

Example 3.5. Consider the field extension $\mathbb{Q}(\zeta_8)/\mathbb{Q}$ where ζ_8 is a primitive 8th root of unity. We have that

$$\zeta_8 = \frac{\sqrt{2}}{2} + i\frac{\sqrt{2}}{2}$$

and $\mathbb{Q}(\zeta_8)/\mathbb{Q}$ is the same field extension as $\mathbb{Q}(i, \sqrt{2})/\mathbb{Q}$. We have

$$[\mathbb{Q}(i, \sqrt{2}) : \mathbb{Q}] = [\mathbb{Q}(i, \sqrt{2}) : \mathbb{Q}(\sqrt{2})][\mathbb{Q}(\sqrt{2}) : \mathbb{Q}] = 2 \cdot 2 = 4.$$

Recall that an algebraic extension is a field extension where every element is algebraic. The result below describes families of algebraic extensions.

Theorem 3.5. *If E is a finite extension of F , then E is an algebraic extension of F .*

Proof. Let $\alpha \in E$ with degree $[E : F] = n$. Then $1, \alpha, \dots, \alpha^n$ are $n+1$ elements while the dimension is n , so they must be linearly dependent, say

$$a_0 + a_1\alpha + \dots + a_n\alpha^n = 0, \quad a_i \in F.$$

Take $p(X) = a_0 + a_1X + \dots + a_nX^n \in F[X]$, α is a root of $p(X)$ and by definition α is algebraic over F . \square

Examples 3.6. 1. By definition, a number field is a finite extension of \mathbb{Q} . Thus a number field is an algebraic extension of \mathbb{Q} .

2. The converse is not true. There are infinite algebraic extensions, for example, the field of all algebraic numbers over the rationals is algebraic and of infinite degree.

3.2 Splitting fields and algebraic closures

For $\alpha \in E$, an extension of F , we have introduced above $F(\alpha)$ as the intersection of all the subfields of E containing F and α . This can be of course generalized if we pick $\alpha_1, \dots, \alpha_k \in E$, and $F(\alpha_1, \dots, \alpha_k)$ is the intersection of all the subfields of E containing F and $\alpha_1, \dots, \alpha_k$.

Definition 3.5. If E is an extension of F and $f \in F[X]$, we say that f **splits** over E if f can be written as $\lambda(X - \alpha_1) \cdots (X - \alpha_k)$ for some $\alpha_1, \dots, \alpha_k \in E$ and $\lambda \in F$.

Definition 3.6. If K is an extension of F and $f \in F[X]$, we say that K is a **splitting field** for f over F if f splits over K but not over any proper subfield of K containing F .

Example 3.7. Consider the polynomial $f(X) = X^3 - 2$ over \mathbb{Q} . Its roots are

$$\sqrt[3]{2}, \sqrt[3]{2} \left(-\frac{1}{2} + i\frac{1}{2}\sqrt{3} \right), \sqrt[3]{2} \left(-\frac{1}{2} - i\frac{1}{2}\sqrt{3} \right).$$

Alternatively, if ζ_3 denotes a primitive 3rd root of unity, we can write the roots as

$$\sqrt[3]{2}, \zeta_3 \sqrt[3]{2}, \zeta_3^2 \sqrt[3]{2}.$$

The polynomial f is irreducible (for example using Eisenstein's criterion). Since it is also monic, it is the minimal polynomial of $\sqrt[3]{2}$, and

$$[\mathbb{Q}(\sqrt[3]{2}) : \mathbb{Q}] = 3.$$

Now since $\sqrt[3]{2}$ and $i\sqrt{3}$ (or ζ_3) generate all the roots of f , the splitting field of f is

$$K = \mathbb{Q}(\sqrt[3]{2}, i\sqrt{3}) = \mathbb{Q}(\sqrt[3]{2}, \zeta_3).$$

We finish by computing the degree of K over \mathbb{Q} . Clearly $i\sqrt{3}$ cannot belong to $\mathbb{Q}(\sqrt[3]{2})$ which is a subfield of \mathbb{R} , thus $[\mathbb{Q}(\sqrt[3]{2}, i\sqrt{3}) : \mathbb{Q}(\sqrt[3]{2})]$ is at least 2. Since $i\sqrt{3}$ is a root of $X^2 + 3 \in \mathbb{Q}(\sqrt[3]{2})[X]$, this degree is exactly 2. By multiplicativity of the degrees, we get that

$$[K : \mathbb{Q}] = 6.$$

Using that ζ_3 is a root of $X^2 + X + 1$ stays irreducible over $\mathbb{Q}(\sqrt[3]{2})$ gives the same result.

Equivalently, K is a splitting field for f over F if f splits over K and K is generated over F by the roots $\alpha_1, \dots, \alpha_k$ of f , that is $K = F(\alpha_1, \dots, \alpha_k)$.

If $f \in F[X]$ and f splits over the extension E of F , then E contains a unique splitting field for f , namely $F(\alpha_1, \dots, \alpha_k)$.

Here is a result on the degree of splitting fields. Note that the above example shows that this bound is tight.

Proposition 3.6. *If $f \in F[X]$ and $\deg f = n$, then f has a splitting field K over F with $[K : F] \leq n!$.*

Proof. First we may assume that $n \geq 1$, for if $n = 0$, then f is constant, and we take $K = F$ with $[K : F] = 1$.

Thus f has at least one root α_1 , and by Theorem 3.1, there is an extension E_1 of F containing α_1 . Since $f(\alpha_1) = 0$, the minimal polynomial $m_1(X)$ of α_1 divides $f(X)$, that is $f(X) = m_1(X)f'(X)$ for some $f'(X)$, and since $\deg f = n$, $\deg m_1(X) \leq n$, implying that $F(\alpha_1)/F$ has degree at most n .

We may then further write $f(X) = (X - \alpha_1)^{r_1}g(X)$ where $g(\alpha_1) \neq 0$ and $\deg g \leq n - 1$. If g is constant, then $f(X)$ has no other root than α_1 , and its splitting field is $F(\alpha_1)/F$ whose degree is at most n which is indeed smaller than $n!$.

Now if g is non-constant, we can iterate on g the reasoning we did on f . Namely, we have that g has degree at least 1, and thus it has at least one root α_2 . Invoking again Theorem 3.1, there is an extension of $F(\alpha_1)$ containing α_2 and the extension $F(\alpha_1, \alpha_2)$ has degree at most $n - 1$ over $F(\alpha_1)$ (corresponding to the case where $r_1 = 1$). Thus we have

$$\begin{aligned} [F(\alpha_1, \alpha_2) : F] &= [F(\alpha_1, \alpha_2) : F(\alpha_1)][F(\alpha_1) : F] \\ &\leq (n - 1)n. \end{aligned}$$

We can now continue inductively to reach that if $\alpha_1, \dots, \alpha_n$ are all the roots of f , then

$$[F(\alpha_1, \alpha_2, \dots, \alpha_n) : F] \leq n!.$$

□

If $f \in F[X]$ and f splits over E , then we may take any root α of f and adjoin it to F to get the extension $F(\alpha)$. More precisely:

Theorem 3.7. *If α and β are roots of the irreducible polynomial $f \in F[X]$ in an extension E of F , then $F(\alpha)$ is isomorphic to $F(\beta)$.*

Proof. If f is not monic, start by dividing f by its leading coefficient, so that we can assume that f is monic. Since f is monic, irreducible and $f(\alpha) = f(\beta) = 0$, f is the minimal polynomial of α and β , say of degree n . Now if $a \in F(\alpha)$, then a can be uniquely written as

$$a = a_0 + a_1\alpha + \dots + a_{n-1}\alpha^{n-1}.$$

The map

$$a_0 + a_1\alpha + \dots + a_{n-1}\alpha^{n-1} \mapsto a_0 + a_1\beta + \dots + a_{n-1}\beta^{n-1}$$

defines a field isomorphism between $F(\alpha)$ and $F(\beta)$. □

When discussing field isomorphisms, one may want to emphasize the base field.

Definition 3.7. If E and E' are extensions of F , and $\iota : E \rightarrow E'$ is an isomorphism, we say that ι is an F -isomorphism if ι fixes F , that is, if

$$\iota(a) = a, \quad a \in F.$$

Given a polynomial $f \in F[X]$, we have discussed its splitting field, namely the smallest field over which f splits. If F is \mathbb{Q} , \mathbb{R} or more generally \mathbb{C} , not only we can find a splitting field for each polynomial, but we know that there is a field C with the property that any polynomial in $\mathbb{C}[X]$ splits over C , namely $C = \mathbb{C}$ itself.

We now would like to express this property in general, without having to assume that F is \mathbb{Q} , \mathbb{R} or \mathbb{C} . Namely, for a general field F , we want an extension C of F such that any polynomial in $C[X]$ splits over C . We will later on add the requirement that this extension is algebraic.

Proposition 3.8. *If C is a field, the following conditions are equivalent.*

1. *Every non-constant polynomial $f \in C[X]$ has at least one root in C .*
2. *Every non-constant polynomial $f \in C[X]$ splits over C .*
3. *Every irreducible polynomial $f \in C[X]$ is linear.*

4. C has no proper algebraic extension.

Proof. We prove $1. \Rightarrow 2. \Rightarrow 3. \Rightarrow 4. \Rightarrow 1.$

1. \Rightarrow 2. Take $f \in C[X]$ a non-constant polynomial. Since f has at least one root, we write $f = (X - \alpha_1)g$ for g some polynomial in $C[X]$. If g is constant, we are done since f splits. If g is non-constant, then again by assumption it has one root and $g = (X - \alpha_2)h$ for some h . We conclude by repeating inductively.
2. \Rightarrow 3. Take $f \in C[X]$ which is irreducible, thus non-constant. By assumption it is a product of linear factors. But f is irreducible, so there can be only one such factor.
3. \Rightarrow 4. Let E be an algebraic extension of C . Take $\alpha \in E$ with minimal polynomial f over C . Then f is irreducible and of the form $X - \alpha \in C[X]$ by assumption. Thus $\alpha \in C$ and $E = C$.
4. \Rightarrow 1. Let f be a non-constant polynomial in $C[X]$, with root α . We can adjoin α to C to obtain $C(\alpha)$. But by assumption, there is no proper algebraic extension of C , so $C(\alpha) = C$ and $\alpha \in C$. Thus f has at least one root in C and we are done.

□

Definition 3.8. A field C as described in the above equivalent properties is said to be **algebraically closed**.

- Examples 3.8.**
1. The field \mathbb{R} is not algebraically closed, since $X^2 + 1 = 0$ has not root in \mathbb{R} .
 2. No finite field \mathbb{F} is algebraically closed, since if a_1, \dots, a_n are all the elements of F , then the polynomial $(X - a_1) \dots (X - a_n) + 1$ has no zero in \mathbb{F} .
 3. The field \mathbb{C} is algebraically closed, this is the fundamental theorem of algebra.
 4. The field of all algebraic numbers is algebraically closed. (We will not prove this here, but for a proof that algebraic numbers in a field extension indeed form a field, see Corollary 3.11 below.)

We can embed an arbitrary field F in an algebraically closed field as follows.

Definition 3.9. An extension C of F is called an **algebraic closure** if C is algebraic over F and C is algebraically closed.

Examples 3.9. To get examples of algebraic closures, we thus need to start with known algebraically closed fields.

1. The field \mathbb{C} is the algebraic closure of \mathbb{R} .

2. The field of all algebraic numbers is the algebraic closure of \mathbb{Q} .

Note that C is minimal among algebraically closed extensions of F . Indeed, let us assume that there is an algebraically closed field K such that $C/K/F$. Let $\alpha \in C$ but $\alpha \notin K$ (it exists if we assume that $C \neq K$). Then α is algebraic over F , and consequently algebraic over K . But since $\alpha \notin K$, the minimal polynomial of α over K cannot contain the factor $X - \alpha$, which contradicts that K is an algebraically closed field.

We can prove the following theorems (we will omit the proof).

Theorem 3.9. 1. Every field F has an algebraic closure.

2. Any two algebraic closures C and C' of F are F -isomorphic.
3. If E is an algebraic extension of F , C is an algebraic closure of F , and ι is an embedding of F into C . Then ι can be extended to an embedding of E into C .

Let us now prove the first transitivity property of field extensions. Several will follow later on in this chapter.

Proposition 3.10. 1. If E is generated over F by finitely many elements $\alpha_1, \dots, \alpha_n$ algebraic over F , then E is a finite extension of F .

2. **(Transitivity of algebraic extensions).** If E is algebraic over K , and K is algebraic over F , then E is algebraic over F .

Proof. 1. Set $E_0 = F$, $E_k = F(\alpha_1, \dots, \alpha_k)$, $1 \leq k \leq n$, in particular $E_n = F(\alpha_1, \dots, \alpha_n) = E$ by definition of E . Then $E_k = E_{k-1}(\alpha_k)$, where α_k is algebraic over F , and hence over E_{k-1} . Now $[E_k : E_{k-1}]$ is the degree of the minimal polynomial of α_k over E_{k-1} , which is finite. By multiplicativity of the degrees, we conclude that

$$[E : F] = \prod_{k=1}^n [E_k : E_{k-1}] < \infty.$$

2. Let $\alpha \in E$ with minimal polynomial

$$m(X) = b_0 + b_1X + \dots + b_{n-1}X^{n-1} + X^n$$

over K since by assumption α is algebraic over K . The coefficients b_i are in K and thus are algebraic over F . Set $L = F(b_0, b_1, \dots, b_{n-1})$, by the first part, L is a finite extension of F . Therefore $m(X) \in L[X]$, α is algebraic over L , and $L(\alpha)$ is a finite extension of L . This gives us the following tower of field extensions:

$$L(\alpha)/L = F(b_0, b_1, \dots, b_{n-1})/F.$$

By transitivity of the degrees, since $[L : F] < \infty$ and $[L(\alpha) : L] < \infty$, we get that $[L(\alpha) : F] < \infty$. We conclude since we know that all finite extensions are algebraic, and thus α is algebraic over F . □

Corollary 3.11. *If E is an extension of F and A is the set of all elements in E that are algebraic over F , then A is a subfield of E .*

Proof. If $\alpha, \beta \in A$, then the sum, difference, product and quotient (if $\beta \neq 0$) of α and β belong to $F(\alpha, \beta)$, which is a finite extension of F by the first part of the above proposition. This is thus an algebraic extension since all finite extensions are, and thus $\alpha + \beta$, $\alpha - \beta$, $\alpha\beta$ and α/β are in A , proving that A is a field. \square

3.3 Separability

If f is a polynomial in $F[X]$, we have seen above that we can construct a splitting field K for f over F , and K is such that all roots of f lie in it. We can thus study the multiplicity of the roots of f in K .

Definition 3.10. An irreducible polynomial $f \in F[X]$ is **separable** if f has no repeated roots in a splitting field. It is called **inseparable** otherwise. Note that if f is not necessarily irreducible, then we call f separable if each of its irreducible factors is separable.

For example $f(X) = (X - 1)^2(X - 2) \in \mathbb{Q}$ is separable, since its irreducible factors $X - 1$ and $X - 2$ are separable.

We start by computing a criterion to test if a polynomial has multiple roots.

Proposition 3.12. *Consider*

$$f(X) = a_0 + a_1X + \cdots + a_nX^n \in F[X]$$

and its formal derivative

$$f'(X) = a_1 + 2a_2X + \cdots + na_nX^{n-1}.$$

Then f has a repeated root in a splitting field if and only if the degree of the greatest common divisor of f and f' is at least 1.

Proof. Let us assume that f has a repeated root in its splitting field, say α . Then we can write

$$f(X) = (X - \alpha)^r h(X)$$

where $r \geq 2$ since we consider a repeated root. Now we compute the derivative of f :

$$f'(X) = r(X - \alpha)^{r-1}h(X) + (X - \alpha)^r h'(X)$$

and since $r - 1 \geq 1$, we have that $(X - \alpha)$ is a factor of both f and f' .

Conversely, let us assume that the greatest common divisor g of f and f' has degree at least 1, and let α be a root of g (in a splitting field). By definition of g , $X - \alpha$ is then a factor of both f and f' . We are left to prove that α is a repeated root of f . Indeed, if it were not the case, then $f(X)$ would be of the form $f(X) = (X - \alpha)h(X)$ where $h(\alpha) \neq 0$ and by computing the derivative, we would get (put $r = 1$ in the above expression for f') $f'(\alpha) = h(\alpha) \neq 0$ which contradicts the fact that $X - \alpha$ is a factor of f' . \square

As a corollary of this result, we can exhibit two classes of separable polynomials.

Corollary 3.13. 1. Over a field of characteristic zero, every polynomial is separable.

2. Over a field F of prime characteristic p , an irreducible polynomial f is inseparable if and only if f' is the zero polynomial (equivalently f is in $F[X^p]$).

Proof. 1. Without loss of generality, consider f an irreducible polynomial in $F[X]$, where F is of characteristic zero. If f is a polynomial of degree n , then its derivative f' is of degree less than n , and it cannot possibly be the zero polynomial. Since f is irreducible, the greatest common divisor of f and f' is either 1 or f , but it cannot be f since f' is of smaller degree. Thus it is 1, and f is separable by the above proposition.

2. We now consider the case where F is of characteristic p . As above, we take f an irreducible polynomial of degree n in $F[X]$ and compute its derivative f' . If f' is non-zero, we can use the same argument. But f' could also be zero, in which case the greatest common divisor of f and f' is actually f , and by the above proposition, f has a multiple root and is then not separable. That $f' = 0$ means that $f \in F[X^p]$ since we work in characteristic p . □

Example 3.10. Polynomials over $\mathbb{R}[X]$ and $\mathbb{Q}[X]$ are separable.

Another class of separable polynomials are polynomials over finite fields, but this asks a little bit more work.

Lemma 3.14. Let F be a finite field of characteristic p . Consider the map

$$f : F \rightarrow F, f(\alpha) = \alpha^p.$$

Then f is an automorphism (called the *Frobenius Automorphism*). In particular, we have for all $\alpha \in F$ that

$$\alpha = \beta^p$$

for some $\beta \in F$.

Proof. We have that f is a ring automorphism since

$$\begin{aligned} f(1) &= 1 \\ f(\alpha + \beta) &= (\alpha + \beta)^p = \alpha^p + \beta^p = f(\alpha) + f(\beta) \\ f(\alpha\beta) &= (\alpha\beta)^p = \alpha^p\beta^p = f(\alpha)f(\beta). \end{aligned}$$

The second set of equalities uses the binomial expansion modulo p . Now f is a monomorphism since F is a field, and an injective map from a finite set to itself is necessarily surjective. □

Proposition 3.15. *Every polynomial is separable over a finite field F (of prime characteristic).*

Proof. Suppose that f is an irreducible polynomial which, by contradiction, has multiple roots in a splitting field. Using the criterion of the corollary, $f(X)$ must be in $F[X^p]$, namely

$$f(X) = a_0 + a_1X^p + \cdots + a_nX^{np}, \quad a_i \in F.$$

Using the bijectivity of the Frobenius automorphism, we can write $a_i = b_i^p$, yielding

$$(b_0 + b_1X + \cdots + b_nX^n)^p = b_0^p + b_1^pX^p + \cdots + b_n^pX^{np} = f(X)$$

which contradicts the irreducibility of f . □

Definition 3.11. If E is an extension of F and $\alpha \in E$, then α is said to be **separable** over F if α is algebraic over F and its minimal polynomial $\mu_{\alpha,F}$ is a separable polynomial. If every element of E is separable over F , we say that E is a **separable extension** of F or that E/F is separable.

Examples 3.11. 1. Typical examples of separable extensions are finite fields and number fields.

2. If F is a field with algebraic closure C , then C contains a smallest field containing all finite separable extensions of F , called the **separable closure** of F . It is a separable extension of F .

Here is a first result on how separability behaves in a tower of extensions.

Lemma 3.16. *If $E/K/F$ and E is separable over F , then K is separable over F and E is separable over K .*

Proof. **K/F is separable.** Since K is a subfield of E , every element $\beta \in K$ belongs to E , and every element of E is separable over F by assumption.

E/K is separable. Take $\alpha \in E$. Since E is separable over F , it is in particular algebraic over F and we may consider the minimal polynomial $\mu_{\alpha,F}$ of α over F . Denote by $\mu_{\alpha,K}$ the minimal polynomial of α over K , we have

$$\mu_{\alpha,K} \mid \mu_{\alpha,F}.$$

Since $\mu_{\alpha,F}$ has no repeated root, neither has $\mu_{\alpha,K}$, and E/K is separable. □

The converse is also true, and gives the transitivity of separable extensions: If K/F and E/K are separable, then E/F is separable.

It is less easy to construct inseparable extensions, but here is a classical example.

Example 3.12. Let \mathbb{F}_p denote the finite field of integers modulo p . Consider the field $F = \mathbb{F}_p(t)$ of rational functions in t with coefficients in the finite field with p elements \mathbb{F}_p . We get a field extension of E/F by adjoining to F a root of the polynomial $X^p - t$ (one has to check that $X^p - t$ is irreducible over $\mathbb{F}_p[t]$). The extension E/F is inseparable since

$$X^p - t = X^p - (\sqrt[p]{t})^p = (X - \sqrt[p]{t})^p,$$

which has multiple roots.

Let E/F be a separable extension of F and let C be an algebraic closure of E . We next count the number of embeddings of E in C that fix F , that is, the number of F -monomorphisms of E into C . We start with a lemma.

Lemma 3.17. *Let $\sigma : E \rightarrow E$ be an F -monomorphism and assume that $f \in F[X]$ splits over E . Then σ permutes the roots of f , namely, if α is a root of f in E then so is $\sigma(\alpha)$.*

Proof. Write $f(X)$ as

$$f(X) = b_0 + b_1X + \cdots + b_nX^n, \quad b_i \in F.$$

If α is a root of f in E , then

$$f(\alpha) = b_0 + b_1\alpha + \cdots + b_n\alpha^n = 0.$$

Apply σ to the above equation, and use that σ is a field homomorphism that fixes F to get

$$b_0 + b_1\sigma(\alpha) + \cdots + b_n\sigma(\alpha)^n = 0,$$

showing that $\sigma(\alpha)$ is a root. \square

Theorem 3.18. *Let E/F be a finite separable extension of degree n , and let σ be an embedding of F into an algebraic closure C . Then σ extends to exactly n embeddings of E in C . Namely, there are exactly n embeddings τ of E into C , such that the restriction $\tau|_F$ of τ to F coincides with σ . In particular, taking σ to be the identity on F , there are exactly n F -monomorphisms of E into C .*

Proof. We do a proof by induction. If $n = 1$, then $E = F$ and σ extends to exactly 1 embedding, namely itself.

We now assume that $n > 1$ and choose $\alpha \in E$, $\alpha \notin F$. Let $f = \mu_{\alpha, F}$ be the minimal polynomial of α over F of degree say r . It is irreducible and separable (E/F is separable by assumption). In order to use the induction hypothesis, we need to split the field extension E/F , which we do by considering the field extension $F(\alpha)$, which satisfies

$$E/F(\alpha)/F, [E : F(\alpha)] = n/r, [F(\alpha) : F] = r.$$

We first take care of the extension $F(\alpha)/F$. Let σ be an embedding of F into C , and define the polynomial $g = \sigma(f)$, where σ is applied on all the coefficients

of f . The polynomial g inherits the property of being irreducible and separable from f . Let β denotes a root of g . We can thus define a unique isomorphism

$$F(\alpha) \rightarrow (\sigma(F))(\beta), \quad b_0 + b_1\alpha + \dots + b_r\alpha^r \mapsto \sigma(b_0) + \sigma(b_1)\beta + \dots + \sigma(b_r)\beta^r$$

and restricted to F it indeed coincides with σ . This isomorphism is defined by the choice of β , and there are exactly r choices for it, corresponding to the r roots of g (note that this is here that the separability of g is crucial). For each of these r isomorphisms, using the induction hypothesis on $[E : F(\alpha)] = n/r < n$, we can extend them to exactly n/r embeddings of E into C . This gives us a total of $n/r \cdot r$ distinct embeddings of E into C extending σ . We conclude by noting that we cannot have more than n such embeddings. \square

We conclude by giving a nice description of finite separable field extensions.

Theorem 3.19. (Theorem of the Primitive Element). *If E/F is a finite separable extension, then*

$$E = F(\gamma)$$

for some $\gamma \in E$. We say that γ is a *primitive element* of E over F .

Proof. Since we have not studied finite fields yet, let us assume that F is an infinite field. (If you have already studied finite fields, then you know we can take γ to be any generator of the cyclic group E^\times).

We proceed by induction on the degree n of the extension E/F . If $n = 1$, then $E = F$ and we can take any element for α .

Let us thus assume $n > 1$, the assumption true up to $n - 1$, and say the degree of E/F is n . Choose $\alpha \in E$ but not in F . We now look at the field extension $E/F(\alpha)$. By induction hypothesis, there is a primitive element β such that

$$E = F(\alpha, \beta).$$

We are now going to prove that there exists a $c \in F$ such that

$$E = F(\alpha + c\beta),$$

that is

$$\gamma = \alpha + c\beta$$

will be the primitive element. We will show that it is enough to take $c \notin S$, where S is a finite subset of F defined as follows: let f be the minimal polynomial of α over F , and let g be the minimal polynomial of β over F , the exceptional set S consists of all $c \in F$ such that

$$c = \frac{\alpha' - \alpha}{\beta - \beta'}$$

for α' a root of f and β' a conjugate of β (we extend $F(\alpha, \beta)$ to a field L in which f and g both split to be able to speak of all their roots).

To show that γ is primitive for $c \notin S$, it is enough to prove that $F(\alpha + c\beta)$ contains β and $\alpha = \gamma - c\beta$ (clearly the reverse inclusion holds: $F(\alpha + c\beta) \subseteq F(\alpha, \beta)$). To this end, it is enough to show that the minimal polynomial of β over $F(\gamma)$ cannot have degree greater or equal to 2, implying that β is in $F(\gamma)$.

Note first that if we take the polynomial $h(X)$ defined by

$$h(X) = f(\gamma - cX) \in F(\gamma)[X]$$

and evaluate it in β , we get

$$h(\beta) = f(\gamma - c\beta) = f(\alpha + c\beta - c\beta) = 0.$$

Thus β is a root of h and the minimal polynomial of β over $F(\gamma)$ divides both g and h , so we are done if we show that the greatest common divisor of g and h in $F(\gamma)[X]$ cannot have degree greater or equal to 2.

Suppose the greatest common divisor does have degree ≥ 2 . Then g and h have as common root in L not only β , but also $\beta' \neq \beta$ in L . This is where we use the separability of g , since otherwise β could be a root with multiplicity 2. Then

$$f(\gamma - c\beta') = 0 \Rightarrow \gamma - c\beta' = \alpha'$$

for some root α' of f , which can be rewritten as

$$\alpha + c\beta - c\beta' = 0 \Rightarrow c = \frac{\alpha' - \alpha}{\beta - \beta'}$$

which is exactly what was ruled out by choosing $c \notin S$. □

Definition 3.12. A **simple extension** is a field extension which is generated by the adjunction of a single element.

Thus the primitive element Theorem above provides a characterization of the finite extensions which are simple.

Example 3.13. Number fields are simple extensions.

3.4 Normality

So far, we have considered two properties of field extensions (both of them being transitive): being algebraic and separable. We now introduce a third property, which is not transitive, the one of being normal.

Definition 3.13. An algebraic extension E/F is **normal** if every irreducible polynomial over F that has at least one root in E splits over E . If we call the other roots of this polynomial the **conjugates** of α , we can rephrase the definition by saying that if $\alpha \in E$, then all conjugates of α over F are in E .

Note that this definition assumes that we start with an algebraic extension.

Example 3.14. Consider the field extension $E = \mathbb{Q}(\sqrt[3]{2})/\mathbb{Q}$. The roots of the irreducible polynomial $f(X) = X^3 - 2$ are

$$\sqrt[3]{2}, \zeta_3 \sqrt[3]{2}, \zeta_3^2 \sqrt[3]{2},$$

where ζ_3 is a primitive 3rd root of unity (for example $\zeta_3 = e^{2\pi i/3}$). Thus E is not a normal extension.

We can give another characterization in terms of monomorphisms of E .

Theorem 3.20. *The finite extension E/F is normal if and only if every F -monomorphism of E into an algebraic closure C is actually an F -automorphism of E . (Finite could be replaced by algebraic, which we will not prove).*

Proof. If E/F is normal, then an F -monomorphism of E into C must map each element of E to one of its conjugates (as is the case in the proof of Lemma 3.17). Thus $\tau(E) \subseteq E$, but $\tau(E)$ is an isomorphic copy of E and thus has the same degree as E and $E = \tau(E)$, showing that τ is indeed an F -automorphism of E .

Conversely, consider $\alpha \in E$ and let β be a conjugate of α over F . There exists an F -monomorphism of E into C that carries α to β (the construction is given in the proof of Theorem 3.18). If all such embeddings are F -automorphisms of E , that means β must be in E , and we conclude that E/F is normal. \square

Here is another characterization of normal extensions in terms of splitting fields.

Theorem 3.21. *The finite extension E/F is normal if and only if E is a splitting field for some polynomial f in $F[X]$.*

Proof. Let E/F be a finite normal extension of degree n , and let $\alpha_1, \dots, \alpha_n$ be a basis for E over F . Consider for each α_i its minimal polynomial f_i over F . By definition of normal extension, since f_i has a root in E , then f_i splits over E , and so does the polynomial

$$f = f_1 \cdots f_n.$$

To prove that E is a splitting field, we are left to prove it is the smallest field over which f splits. This is here that we understand why we take such an f . If f were to split over a subfield K , that is K such that

$$F \subset K \subset E$$

then each $\alpha_i \in K$, and $K = E$ (this is a conclusion we cannot reach if we take for f only one f_i or a subset of them). This proves that E is a splitting field for f over F .

Conversely, let E be a splitting field for some f over F , whose roots are denoted by $\alpha_1, \dots, \alpha_n$. Let τ be an F -monomorphism of E into an algebraic closure, that is τ takes each α_i into another root of f .

Since E is a splitting field for f , we have

$$F(\alpha_1, \dots, \alpha_n) = E$$

and $\tau(E) \subset E$. Thus since E and $\tau(E)$ have same dimension, we get that

$$\tau(E) = E$$

and τ is actually an automorphism of E , and by the above theorem, we conclude the E/F is normal. \square

As a corollary, we see how a subextension inherits the property of normality.

Corollary 3.22. *Let $E/K/F$ be a finite extension ($[E : F] < \infty$). If E/F is normal, so is E/K .*

Proof. Since E/F is normal, E is a splitting field for some polynomial $f \in F[X]$, that is E is generated over F by the roots of f . Since $f \in F[X] \subset K[X]$, f can also be seen as a polynomial in $K[X]$ and E is generated over K by the roots of f , and again by the above theorem, E/K is normal. \square

There is no reason for an arbitrary field extension E/F to be normal. However, if E/F is finite (or more generally algebraic) one can always embed it in normal extension.

Definition 3.14. Let E/F be an algebraic extension. The **normal closure** of E/F is an extension field N of E such that N/E is normal and N is minimal with this property.

If E/F is finite, we can see it as follows: E is finitely generated over F , so it can be written as $E = F(\alpha_1, \dots, \alpha_n)$. Let now K be a normal extension of F that contains E :

$$K/E/F.$$

Since K is normal, it must contain not only all the α_i but also all their conjugates. Let f_i be the minimal polynomial of α_i , $i = 1, \dots, n$. Then we can rephrase the last statement and say that K must contain all the roots of f_i , $i = 1, \dots, n$. Consider the polynomial

$$f = f_1 \cdots f_n.$$

Then K must contain the splitting field N for f over F . But N/F is normal, so N must be the smallest normal extension of F that contains E . Thus N is a normal closure of E over F .

The main definitions and results of this chapter are

- **(3.1).** Definitions of: field extension, minimal polynomial, degree of a field extension, field homomorphism, algebraic, transcendental. That the degree is multiplicative.
- **(3.2).** Definitions of: to split, splitting field, algebraically closed, algebraic closure. Transitivity of algebraic extensions.
- **(3.3).** Definition of separability, typical separable extensions, separability in extension towers, number of embeddings into an algebraic closure, primitive element Theorem.
- **(3.4).** Definition of normality, two equivalent characterizations of normal extensions.

Chapter 4

Galois Theory

Galois theory is named after the French mathematician Evariste Galois.

Galois was born in 1811, and had what could be called the life of a misunderstood genius. At the age of 15, he was already reading material written for professional mathematicians. He took the examination to the “Ecole Polytechnique” to study mathematics but failed and entered the “Ecole Normale” in 1828. He wrote his first paper at the age of 18. He tried to advertise his work, and sent his discoveries in the theory of polynomial equations to the Academy of Sciences, where Cauchy rejected his memoir. He did not get discouraged, and in 1830, he wrote again his researches to the Academy of Sciences, where this time Fourier got his manuscript. However, Fourier died before reading it.

A year later, he made a third attempt, and sent to the Academy of Sciences a memoir called “On the conditions of solvability of equations by radicals”. Poisson was a referee, and he answered several months later, declaring the paper incomprehensible.

In 1832, he got involved in a love affair, but got rejected due to a rival, who challenged him to a duel. The night before the duel, he wrote a letter to his friend, where he reported his mathematical discoveries. He died during the duel with pistols in 1832.

It is after his death that his friend insisted to have his letter published, which was finally done by the mathematician Chevalier.

4.1 Galois group and fixed fields

Definition 4.1. If E/F is normal and separable, it is said to be a [Galois extension](#), or alternatively, we say that E is Galois over F .

Take E/F a Galois extension of degree n . Since it is separable of degree n , we know that there are exactly n F -monomorphisms of E into an algebraic closure C . But E/F being also normal, every F -automorphism into C is actually and



Figure 4.1: Evariste Galois (1811-1832)

F -automorphism of E . Thus there are exactly $n = [E : F]$ F -automorphisms of E .

We can define the notion of a Galois group for an arbitrary field extension.

Definition 4.2. If E/F is a field extension, the **Galois group** of E/F , denoted by $\text{Gal}(E/F)$, is the set of F -automorphisms of E . It forms a group under the composition of functions.

Example 4.1. If $E = \mathbb{Q}(\sqrt[3]{2})$, then $\text{Gal}(E/\mathbb{Q}) = \{1\}$, that is the identity on E .

The above example illustrates the fact that though one can always define a Galois group, we need the extension to be actually Galois to say that the order of the Galois group is actually the degree of the field extension.

Definition 4.3. Let $G = \text{Gal}(E/F)$ be the Galois group of the extension E/F . If H is a subgroup of G , the **fixed field** of H is the set of elements fixed by every automorphism in H , that is

$$\mathcal{F}(H) = \{x \in E, \sigma(x) = x \text{ for all } \sigma \in H\}.$$

Vice-versa, if K is an intermediate field, define

$$\mathcal{G}(K) = \text{Gal}(E/K) = \{\sigma \in G, \sigma(x) = x \text{ for all } x \in K\}.$$

It is the group fixing K .

Galois theory has much to do with studying the relations between fixed fields and fixing groups.

Proposition 4.1. Let E/F be a finite Galois extension with Galois group $G = \text{Gal}(E/F)$. Then

1. The fixed field of G is F .

2. If H is a proper subgroup of G , then the fixed field $\mathcal{F}(H)$ of H properly contains F .

Proof. 1. Let F_0 be the fixed field of G (and we have the field extensions $E/F_0/F$). We want to prove that $F_0 = F$.

We first note that if σ is an F -automorphism of E (that is σ is in G), then by definition of F_0 , σ fixes everything in F_0 , meaning that σ is an F_0 -automorphism. Thus the F -automorphisms in the group G coincide with the F_0 -automorphisms in the group G .

Now we further have that E/F_0 is Galois: indeed, we have $E/F_0/F$ with E/F Galois thus normal and separable, and E/F_0 inherits both properties.

We now look at the degrees of the extensions considered:

$$|\text{Gal}(E/F_0)| = [E : F_0], \quad |\text{Gal}(E/F)| = [E : F],$$

since both are Galois. Furthermore by the first remark, the number of F - and F_0 - automorphisms in G coincide:

$$|\text{Gal}(E/F_0)| = |\text{Gal}(E/F)|$$

showing that

$$[E : F_0] = [E : F]$$

and by multiplicativity of the degrees

$$[E : F] = [E : F_0][F_0 : F] \Rightarrow [F_0 : F] = 1$$

and $F = F_0$.

2. In order to prove that $F \subsetneq \mathcal{F}(H)$, let us assume by contradiction that $F = \mathcal{F}(H)$.

Since we consider a finite Galois extension, we can invoke the Theorem of the Primitive Element and claim that

$$E = F(\alpha), \quad \alpha \in E. \tag{4.1}$$

Consider the polynomial

$$f(X) = \prod_{\sigma \in H} (X - \sigma(\alpha)) \in E[X].$$

It is a priori in $E[X]$, but we will prove now that it is actually in $F[X]$. Since by contradiction we are assuming that $F = \mathcal{F}(H)$, it is enough to prove that $f(X)$ is fixed by H . Indeed, take $\tau \in H$, then

$$\prod_{\sigma \in H} (X - \tau\sigma(\alpha)) = \prod_{\sigma \in H} (X - \sigma(\alpha))$$

since $\tau\sigma$ ranges over all H as does σ .

Thus $f(X) \in F[X]$ and $f(\alpha) = 0$ (σ must be the identity once while ranging through H). Now on the one hand, we have

$$\deg f = |H| < |G| = [E : F]$$

since we assume that H is proper and E/F is Galois. On the other hand,

$$\deg f \geq [F(\alpha) : F] = [E : F]$$

since f is a multiple of the minimal polynomial of α over F (equality holds if f is the minimal polynomial of α over F), and $E = F(\alpha)$ by (4.1). We cannot possibly have $\deg f < [E : F]$ and $\deg f \geq [E : F]$ at the same time, which is a contradiction and concludes the proof. \square

4.2 The fundamental Theorem of Galois theory

The most significant discovery of Galois is that (surely not in these terms!) under some hypotheses, there is a one-to-one correspondence between

1. subgroups of the Galois group $\text{Gal}(E/F)$
2. subfields M of E such that $F \subseteq M$.

The correspondence goes as follows:

- To each intermediate subfield M , associate the group $\text{Gal}(E/M)$ of all M -automorphisms of E :

$$\begin{aligned} \mathcal{G} = \text{Gal} : \{\text{intermediate fields}\} &\rightarrow \{\text{subgroups of } \text{Gal}(E/F)\} \\ M &\mapsto \mathcal{G}(M) = \text{Gal}(E/M). \end{aligned}$$

- To each subgroup H of $\text{Gal}(E/F)$, associate the fixed subfield $\mathcal{F}(H)$:

$$\begin{aligned} \mathcal{F} : \{\text{subgroups of } \text{Gal}(E/F)\} &\rightarrow \{\text{intermediate fields}\} \\ H &\mapsto \mathcal{F}(H). \end{aligned}$$

We will prove below that, under the right hypotheses, we actually have a bijection (namely \mathcal{G} is the inverse of \mathcal{F}). Let us start with an example.

Example 4.2. Consider the field extension $E = \mathbb{Q}(i, \sqrt{5})/\mathbb{Q}$. It has four \mathbb{Q} -automorphisms, given by (it is enough to describe their actions on i and $\sqrt{5}$):

$$\begin{aligned} \sigma_1 : \quad i &\mapsto i, & \sqrt{5} &\mapsto \sqrt{5} \\ \sigma_2 : \quad i &\mapsto -i, & \sqrt{5} &\mapsto \sqrt{5} \\ \sigma_3 : \quad i &\mapsto i, & \sqrt{5} &\mapsto -\sqrt{5} \\ \sigma_4 : \quad i &\mapsto -i, & \sqrt{5} &\mapsto -\sqrt{5} \end{aligned}$$

thus

$$\text{Gal}(E/\mathbb{Q}) = \{\sigma_1, \sigma_2, \sigma_3, \sigma_4\}.$$

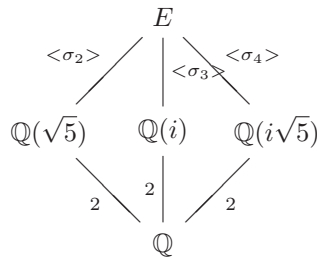
The proper subgroups of $\text{Gal}(E/\mathbb{Q})$ are

$$\{\sigma_1\}, \{\sigma_1, \sigma_2\}, \{\sigma_1, \sigma_3\}, \{\sigma_1, \sigma_4\}$$

and their corresponding subfields are

$$E, \mathbb{Q}(\sqrt{5}), \mathbb{Q}(i), \mathbb{Q}(i\sqrt{5}).$$

We thus get the following diagram:



Theorem 4.2. *Let E/F be a finite Galois extension with Galois group G .*

1. *The map \mathcal{F} is a bijection from subgroups to intermediate fields, with inverse \mathcal{G} .*
2. *Consider the intermediate field $K = \mathcal{F}(H)$ which is fixed by H , and $\sigma \in G$. Then the intermediate field*

$$\sigma K = \{\sigma(x), x \in K\}$$

is fixed by $\sigma H \sigma^{-1}$, namely $\sigma K = \mathcal{F}(\sigma H \sigma^{-1})$.

Proof. 1. We first consider the composition of maps

$$H \rightarrow \mathcal{F}(H) \rightarrow \mathcal{G}\mathcal{F}(H).$$

We need to prove that $\mathcal{G}\mathcal{F}(H) = H$. Take σ in H , then σ fixes $\mathcal{F}(H)$ by definition and $\sigma \in \text{Gal}(E/\mathcal{F}(H)) = \mathcal{G}(\mathcal{F}(H))$, showing that

$$H \subseteq \mathcal{G}\mathcal{F}(H).$$

To prove equality, we need to rule out the strict inclusion. If H were a proper subgroup of $\mathcal{G}(\mathcal{F}(H))$, by the above proposition the fixed field $\mathcal{F}(H)$ of H should properly contain the fixed field of $\mathcal{G}\mathcal{F}(H)$ which is $\mathcal{F}(H)$ itself, a contradiction, showing that

$$H = \mathcal{G}\mathcal{F}(H).$$

Now consider the reverse composition of maps

$$K \rightarrow \mathcal{G}(K) \rightarrow \mathcal{FG}(K).$$

This time we need to prove that $K = \mathcal{FG}(K)$. But

$$\mathcal{FG}(K) = \text{fixed field by } \text{Gal}(E/K)$$

which is exactly K by the above proposition (its first point).

2. It is enough to compute $\mathcal{F}(\sigma H \sigma^{-1})$ and show that it is actually equal to $\sigma K = \sigma \mathcal{F}(H)$.

$$\begin{aligned} \mathcal{F}(\sigma H \sigma^{-1}) &= \{x \in E, \sigma \tau \sigma^{-1}(x) = x \text{ for all } \tau \in H\} \\ &= \{x \in E, \tau \sigma^{-1}(x) = \sigma^{-1}(x) \text{ for all } \tau \in H\} \\ &= \{x \in E, \sigma^{-1}(x) \in \mathcal{F}(H)\} \\ &= \{x \in E, x \in \sigma(\mathcal{F}(H))\} = \sigma(\mathcal{F}(H)). \end{aligned}$$

□

We now look at subextensions of the finite Galois extension E/F and ask about their respective Galois group.

Theorem 4.3. *Let E/F be a finite Galois extension with Galois group G . Let K be an intermediate subfield, fixed by the subgroup H .*

1. *The extension E/K is Galois.*
2. *The extension K/F is normal if and only if H is a normal subgroup of G .*
3. *If H is a normal subgroup of G , then*

$$\text{Gal}(K/F) \simeq G/H = \text{Gal}(E/F)/\text{Gal}(E/K).$$

4. *Whether K/F is normal or not, we have*

$$[K : F] = [G : H].$$

Proof. 1. That E/K is Galois is immediate from the fact that a subextension $E/K/F$ inherits normality and separability from E/F .

2. First note that σ is an F -monomorphism of K into E if and only if σ is the restriction to K of an element of G : if σ is an F -monomorphism of K into E , it can be extended to an F -monomorphism of E into itself thanks to the normality of E . Conversely, if τ is an F -automorphism of E , then $\sigma = \tau|_K$ is surely a F -monomorphism of K into E .

Now, this time by a characterization of a normal extension, we have

$$K/F \text{ normal} \iff \sigma(K) = K \text{ for all } \sigma \in G.$$

Since $K = \mathcal{F}(H)$, we just rewrite

$$K/F \text{ normal} \iff \sigma(\mathcal{F}(H)) = \mathcal{F}(H) \text{ for all } \sigma \in G.$$

Now by the above theorem, we know that $\sigma(\mathcal{F}(H)) = \mathcal{F}(\sigma H \sigma^{-1})$, and we have

$$K/F \text{ normal} \iff \mathcal{F}(\sigma H \sigma^{-1}) = \mathcal{F}(H) \text{ for all } \sigma \in G.$$

We are almost there, we know use again the above theorem that tells us that \mathcal{F} is invertible, with inverse \mathcal{G} , to get the conclusion:

$$K/F \text{ normal} \iff \sigma H \sigma^{-1} = H \text{ for all } \sigma \in G.$$

3. To prove this isomorphism, we will use the 1st isomorphism Theorem for groups. Consider the group homomorphism

$$\text{Gal}(E/F) \rightarrow \text{Gal}(K/F), \sigma \mapsto \sigma|_K.$$

This map is surjective (we showed it above, when we mentioned that we can extend $\sigma|_K$ to σ). Its kernel is given by

$$\text{Ker} = \{\sigma, \sigma|_K = 1\} = H = \text{Gal}(E/K).$$

Applying the 1st isomorphism Theorem for groups, we get

$$\text{Gal}(K/F) \simeq \text{Gal}(E/F)/\text{Gal}(E/K).$$

4. Finally, by multiplicativity of the degrees:

$$[E : F] = [E : K][K : F].$$

Since E/F and E/K are Galois, we can rewrite

$$|G| = |H|[K : F].$$

We conclude by Lagrange Theorem:

$$[G : H] = |G|/|H| = [K : F].$$

□

4.3 Finite fields

We will provide a precise classification of finite fields.

Theorem 4.4. *Let E be a finite field of characteristic p .*

1. *The cardinality of E is*

$$|E| = p^n,$$

for some $n \geq 1$. It is denoted $E = \mathbb{F}_{p^n}$.

2. Furthermore, E is the splitting field for the separable polynomial

$$f(X) = X^{p^n} - X$$

over \mathbb{F}_p , so that any finite field with p^n elements is isomorphic to E . In fact, E coincides with the set of roots of f .

Proof. 1. Let \mathbb{F}_p be the finite field with p elements, given by the integers modulo p . Since E has characteristic p , it contains a copy of \mathbb{F}_p . Thus E is a field extension of \mathbb{F}_p , and we may see E as a vector space over \mathbb{F}_p . If the dimension is n , then let $\alpha_1, \dots, \alpha_n$ be a basis. Every x in E can be written as

$$x = x_1\alpha_1 + \cdots + x_n\alpha_n$$

and there are p choices for each x_i , thus a total of p^n different elements in E .

2. Let E^\times be the multiplicative group of non-zero elements of E . If $\alpha \in E^\times$, then

$$\alpha^{p^n - 1} = 1$$

by Lagrange's Theorem, so that

$$\alpha^{p^n} = \alpha$$

for all α in E (including $\alpha = 0$). Thus each element of E is a root of f , and f is separable.

Now f has at most p^n distinct roots, and we have already identified the p^n elements of E as roots of f .

□

Corollary 4.5. *If E is a finite field of characteristic p , then E/\mathbb{F}_p is a Galois extension, with cyclic Galois group, generated by the Frobenius automorphism*

$$\sigma : x \mapsto \sigma(x) = x^p, \quad x \in E.$$

Proof. By the above proposition, we know that E is a splitting field for a separable polynomial over \mathbb{F}_p , thus E/\mathbb{F}_p is Galois.

Since $x^p = x$ for all x in \mathbb{F}_p , we have that

$$\mathbb{F}_p \subset \mathcal{F}(\langle \sigma \rangle)$$

that is \mathbb{F}_p is contained in the fixed field of the cyclic subgroup generated by the Frobenius automorphism σ . But conversely, each element fixed by σ is a root of $X^p - X$ so $\mathcal{F}(\langle \sigma \rangle)$ has at most p elements. Consequently

$$\mathbb{F}_p = \mathcal{F}(\langle \sigma \rangle)$$

and

$$\text{Gal}(E/\mathbb{F}_p) = \langle \sigma \rangle.$$

□

This can be generalized when the base field is larger than \mathbb{F}_p .

Corollary 4.6. *Let E/F be a finite field extension with $|E| = p^n$ and $|F| = p^m$. Then E/F is a Galois extension and $m|n$. Furthermore, the Galois group is cyclic, generated by the automorphism*

$$\tau : x \mapsto \tau(x) = x^{p^m}, \quad x \in E.$$

Proof. If the degree $[E : F] = d$, then every x in E can be written as

$$x = x_1\alpha_1 + \cdots + x_d\alpha_d$$

and there are p^m choices for each x_i , thus a total of

$$(p^m)^d = p^n$$

different elements in E , so that

$$d = n/m \text{ and } m|n.$$

The same proof as for the above corollary holds for the rest. \square

Thus a way to construct a finite field E is, given p and n , to construct $E = \mathbb{F}_{p^n}$ as a splitting field for $X^{p^n} - X$ over \mathbb{F}_p .

Theorem 4.7. *If G is a finite subgroup of the multiplicative group of an arbitrary field, then G is cyclic. Thus in particular, the multiplicative group E^\times of a finite field E is cyclic.*

Proof. The proof relies on the following fact: if G is a finite abelian group, it contains an element g whose order r is the exponent of G , that is, the least common multiple of the orders of all elements of G .

Assuming this fact, we proceed as follows: if $x \in G$, then its order divides r and thus

$$x^r = 1.$$

Therefore each element of G is a root of $X^r - 1$ and

$$|G| \leq r.$$

Conversely, $|G|$ is a multiple of the order of every element, so $|G|$ is at least as big as their least common multiple, that is

$$|G| \geq r$$

and

$$|G| = r.$$

Since the order of $|G|$ is r , and it coincides with the order of the element g whose order is the exponent, we have that G is generated by g , that is $G = \langle g \rangle$ is cyclic. \square

Since E^\times is cyclic, it is generated by a single element, say α :

$$E = \mathbb{F}_p(\alpha)$$

and α is called a **primitive element** of E . The minimal polynomial of α is called a **primitive polynomial**.

Example 4.3. Consider the following irreducible polynomial

$$g(X) = X^4 + X + 1$$

over \mathbb{F}_2 . Let α be a root of $g(X)$. A direct computation shows that α is primitive:

$$\alpha^0 = 1, \dots, \alpha^4 = \alpha + 1, \dots, \alpha^7 = \alpha^3 + \alpha + 1, \dots, \alpha^{14} = 1 + \alpha^3.$$

4.4 Cyclotomic fields

Definition 4.4. A **cyclotomic extension** of a field F is a splitting field E for the polynomial

$$f(X) = X^n - 1$$

over F . The roots of f are called **n th roots of unity**.

The n th roots of unity form a multiplicative subgroup of the group E^\times of non-zero elements of E , and thus must be cyclic. A **primitive n th root of unity** is an n th root of unity whose order in E^\times is n . It is denoted ζ_n .

From now on, we will assume that we work in a characteristic $\text{char}(F)$ such that $\text{char}(F)$ does not divide n . (Otherwise, we have $n = m\text{char}(F)$ and $0 = \zeta_n^n - 1 = (\zeta^m - 1)^{\text{char}(F)}$ and the order of ζ_n is less than n .)

Example 4.4. The field $\mathbb{Q}(\zeta_p)$ where p is a prime and ζ_p is a primitive p th root of unity is a cyclotomic field over \mathbb{Q} .

Let us look at the Galois group $\text{Gal}(E/F)$ of the cyclotomic extension E/F . Then $\sigma \in \text{Gal}(E/F)$ must map a primitive n th root of unity ζ_n to another primitive n th root of unity ζ_n^r , with $(r, n) = 1$. We can then identify σ with r , and this shows that

$$\text{Gal}(E/F) \simeq U_n$$

where U_n denotes the group of units modulo n . This shows that the Galois group is abelian.

Example 4.5. Consider the field extension $\mathbb{Q}(\zeta_3)/\mathbb{Q}$. We have

$$X^3 - 1 = (X - 1)(X^2 + X + 1).$$

The Galois group is given by:

$$\begin{aligned} \sigma : \zeta_3 &\mapsto \zeta_3^2 \\ \sigma^2 : \zeta_3 &\mapsto \zeta_3 \end{aligned}$$

and the group U_3 of units modulo 3 is $U_3 = \{1, 2\}$. Thus

$$\text{Gal}(\mathbb{Q}(\zeta_3)/\mathbb{Q}) = \{\sigma, 1\} \simeq \{2, 1\} = (\mathbb{Z}/3\mathbb{Z})^\times.$$

Finally, since E/F is Galois (under the above assumption)

$$[E : F] = |\text{Gal}(E/F)| = \varphi(n)$$

where $\varphi(n)$ is the Euler totient function.

From now on, we fix the base field $F = \mathbb{Q}$. This means that a primitive n th root of unity ζ_n is given by

$$\zeta_n = e^{i2\pi r/n}, \quad (r, n) = 1.$$

Definition 4.5. The n th cyclotomic polynomial is defined by

$$\Psi_n(X) = \prod_{(i,n)=1} (X - \zeta_n^i),$$

where the product is taken over all primitive n th roots of unity in \mathbb{C} .

The degree of $\Psi_n(X)$ is thus

$$\deg(\Psi_n) = \varphi(n).$$

Example 4.6. For $n = 1, 2$, we have

$$\Psi_1(X) = X - 1, \quad \Psi_2(X) = X - (-1) = X + 1.$$

Computing a cyclotomic polynomial is not that easy. Here is a formula that can help.

Proposition 4.8. *We have*

$$X^n - 1 = \prod_{d|n} \Psi_d(X).$$

In particular, if $n = p$ a prime, then d is either 1 or p and

$$X^p - 1 = \Psi_1(X)\Psi_p(X) = (X - 1)\Psi_p(X)$$

from which we get

$$\Psi_p(X) = \frac{X^p - 1}{X - 1} = X^{p-1} + X^{p-2} + \cdots + X + 1.$$

Proof. We prove equality by comparing the roots of both monic polynomials.

If ζ is a n th root of unity, then by definition

$$\zeta^n = 1$$

and its order d divides n . Thus ζ is actually a primitive d th root of unity, and a root of $\Psi_d(X)$.

Conversely, if $d|n$, then any root of $\Psi_d(X)$ is a d th root hence a n th root of unity. \square

Examples 4.7. For $n = 3$ and 5 , we have a prime and thus we can use the above formula:

$$\begin{aligned}\Psi_3(X) &= X^2 + X + 1 \\ \Psi_5(X) &= X^4 + X^3 + X^2 + X + 1.\end{aligned}$$

For $n = 4$ the primitive 4th roots of unity are $\pm i$, and by definition

$$\Psi_4(X) = (X - i)(X + i) = X^2 + 1.$$

Finally for $n = 6$, the possible values for d are 1, 2, 3 and 6. Thus

$$\Psi_6(X) = \frac{X^6 - 1}{(X - 1)(X + 1)(X^2 + X + 1)} = X^2 - X + 1.$$

From the above examples, it is tempting to say that in general $\Psi_n(X)$ has integer coefficients. It happens to be true.

Proposition 4.9. *The n th cyclotomic polynomial $\Psi_n(X)$ satisfies*

$$\Psi_n(X) \in \mathbb{Z}[X].$$

Proof. We proceed by induction on n . It is true for $n = 1$ since $X - 1 \in \mathbb{Z}[X]$. Let us suppose it is true for $\Psi_k(X)$ where k is up to $n - 1$, and prove it is also true for n .

Using the above proposition, we know that

$$\begin{aligned}X^n - 1 &= \prod_{d|n} \Psi_d(X) \\ &= \Psi_n(X) \prod_{d|n, d < n} \Psi_d(X).\end{aligned}$$

The aim is to prove that $\Psi_n(X) \in \mathbb{Z}[X]$:

$$\Psi_n(X) = \frac{X^n - 1}{\prod_{d|n, d < n} \Psi_d(X)}.$$

First note that $\Psi_n(X)$ has to be monic (by definition), and both $X^n - 1$ and $\Psi_d(X)$ (by induction hypothesis) are in $\mathbb{Z}[X]$. We can thus conclude invoking the division algorithm for polynomials in $\mathbb{Z}[X]$. \square

We conclude by proving the irreducibility of the cyclotomic polynomials.

Theorem 4.10. *The cyclotomic polynomial $\Psi_n(X)$ is irreducible over \mathbb{Q} .*

Proof. Let $f(X)$ be the minimal polynomial of ζ_n , a primitive n th root of unity over $\mathbb{Q}(X)$. We first note that by definition $f(X)$ is monic, and thus since $f(X) | X^n - 1$, we have

$$X^n - 1 = f(X)g(X) \tag{4.2}$$

and $f(X)$ and $g(X)$ must be in $\mathbb{Z}[X]$.

To prove that $\Psi_n(x)$ is irreducible, we will actually prove that

$$\Psi_n(X) = f(X).$$

To prove the equality, it is enough to show that every root of $\Psi_n(X)$ is a root of $f(X)$.

We need the following intermediate result: if p does not divide n , then

$$f(\zeta_n^p) = 0.$$

Let us prove this result. Suppose by contradiction that this is not the case, namely $f(\zeta_n^p) \neq 0$. By (4.2), we have

$$X^n - 1 = f(X)g(X),$$

which evaluated in $X = \zeta_n^p$ yields

$$(\zeta_n^p)^n - 1 = 0 = f(\zeta_n^p)g(\zeta_n^p)$$

implying by our assumption that $f(\zeta_n^p) \neq 0$ that

$$g(\zeta_n^p) = 0,$$

or in other words, ζ_n is a root of $g(X^p)$. But by definition of minimal polynomial, we have that $f(X)$ must then divide $g(X^p)$, that is

$$g(X^p) = f(X)h(X), \quad h(X) \in \mathbb{Z}[X].$$

Since $g(X^p)$, $f(X)$ and $h(X)$ are in $\mathbb{Z}[X]$, we can look at their reduction modulo p , that is work in $\mathbb{F}_p[X]$. We will denote $\bar{p}(X)$ the polynomial obtained from $p(X)$ by taking all its coefficients modulo p : if $p(X) = \sum_{i=0}^n a_i X^i$, then $\bar{p}(X) = \sum_{i=0}^n (a_i \bmod p) X^i$. Therefore

$$\bar{g}(X^p) = \bar{f}(X)\bar{h}(X) \in \mathbb{F}_p[X].$$

By working in $\mathbb{F}_p[X]$, we are now allowed to write that

$$\bar{g}(X^p) = \bar{g}(X)^p$$

and thus

$$\bar{g}(X)^p = \bar{f}(X)\bar{h}(X) \in \mathbb{F}_p[X].$$

This tells us that any irreducible factor of $\bar{f}(X)$ divides $\bar{g}(X)$ and consequently \bar{f} and \bar{g} have a common factor. Looking at (4.2) in $\mathbb{F}_p[X]$ gives

$$X^n - \bar{1} = \bar{f}(X)\bar{h}(X) \in \mathbb{F}_p[X].$$

Since \bar{f} and \bar{g} have a common factor, $X^n - \bar{1}$ has a multiple root, which cannot be since we have assumed that p does not divide n . This proves the claim.

To summarize, we have just proven that if p does not divide n , then $f(\zeta_n^p)$ is another root of f . Since all primitive n th roots of unity can be obtained from ζ_n by successive prime powers, we have that all primitive n th roots of unity are actually roots of $f(X)$, and we know that there are $\varphi(n)$ of them, which is also the degree of $\Psi_n(X)$. This concludes the proof, since

$$\deg f(X) \geq \varphi(n) = \deg(\Psi_n(X)) \Rightarrow f(X) = \Psi_n(X).$$

□

4.5 Solvability by radicals

The question of solvability by radicals is the one of solving polynomial equations under the restriction that we are only allowed to perform addition, subtraction, multiplication, division, and taking n th roots.

For example, we know (Fontana-Tartaglia, 1535) that for a cubic equation

$$X^3 + pX = q,$$

the solution is given by

$$X = \sqrt[3]{\frac{q}{2} + \sqrt{\frac{p^3}{27} + \frac{q^2}{4}}} + \sqrt[3]{\frac{q}{2} - \sqrt{\frac{p^3}{27} + \frac{q^2}{4}}}.$$

By the 16th century all polynomial equations of degree smaller or equal to 4 were solved. The natural question was then: what happens with quintic equations? Euler failed to give an answer, Lagrange (1770) proved that it depends on finding functions of the roots which are unchanged by certain permutations of the roots, and that this approach works up to degree 4 and fails for 5. Abel showed (1824) that quintics are insolvable by radicals. The next question thus became: decide whether or not a given equation can be solved by radicals. Liouville (1843) found the answer in Galois's papers.

The answer is to be found by connecting the problem with field theory as follows. We first need to define the notion of a radical extension. Informally, a radical extension is obtained by adjoining a sequence of n th roots. For example, to get a radical extension of \mathbb{Q} containing

$$\sqrt[3]{11} \sqrt[5]{\frac{7 + \sqrt{3}}{2}} + \sqrt[4]{1 + \sqrt[3]{4}},$$

we must adjoin

$$\alpha = \sqrt[3]{11}, \beta = \sqrt{3}, \gamma = \sqrt[5]{\frac{7 + \beta}{2}}, \delta = \sqrt[3]{4}, \epsilon = \sqrt[4]{1 + \delta}.$$

This can be stated formally:

Definition 4.6. An extension E/F is **radical** if $E = F(\alpha_1, \dots, \alpha_n)$ where for all $i = 1, \dots, m$, there exists an integer $n(i)$ such that

$$\alpha_i^{n(i)} \in F(\alpha_1, \dots, \alpha_{i-1}), \quad i \geq 2.$$

The α_i 's are said to form a radical sequence for E/F .

Example 4.8. The expression

$$\sqrt[3]{11} \sqrt[5]{\frac{7 + \sqrt{3}}{2}} + \sqrt[4]{1 + \sqrt[3]{4}}$$

is contained in $\mathbb{Q}(\alpha, \beta, \gamma, \delta, \epsilon)$, where

$$\alpha^3 = 11, \beta^2 = 3, \gamma^5 = \frac{7 + \beta}{2}, \delta^3 = 4, \epsilon^4 = 1 + \delta.$$

Definition 4.7. Let f be a polynomial over a field F of characteristic zero (this is a simplifying assumption). We say that f is **solvable (soluble) by radicals** if there exists a field E containing a splitting field for f such that E/F is a radical extension.

We want to connect radical extensions and solvable groups. Here is the main theorem:

Theorem 4.11. *If F is a field of characteristic zero, and $F \subseteq E \subseteq M$ where M/F is a radical extension, then the Galois group of E/F is a solvable group.*

Thus a solvable (by radicals) polynomial has a solvable Galois group (of a splitting field over the base field).

Recall that a group G is solvable if G has a normal series

$$\{1\} = G_r \trianglelefteq G_{r-1} \trianglelefteq \dots \trianglelefteq G_0 = G$$

with G_i/G_{i+1} abelian. The proof takes some fair amount of work, though the idea is simple. A radical extension is a series of extensions by n th roots. Such extensions have abelian Galois groups (to be proven though...), so the Galois group of a radical extension is made up by fitting together a sequence of abelian groups (unfortunately, the proof is not that simple...)

We can restate the above result in terms of polynomials.

Theorem 4.12. *Let f be a polynomial over a field E of characteristic zero. If f is solvable by radicals then its Galois group (that is the Galois group of its splitting field) over E is a solvable group.*

To find a polynomial which is not solvable by radicals, it suffices to find one whose Galois group is not solvable.

Lemma 4.13. *Let p be a prime, f an irreducible polynomial of degree p over \mathbb{Q} . Suppose that f has precisely two non-real zeros in \mathbb{C} . Then the Galois group of f over \mathbb{Q} is the symmetric group S_p .*

Theorem 4.14. *The polynomial $X^5 - 6X + 3$ over \mathbb{Q} is not solvable by radicals.*

The proof consists of showing that the polynomial is irreducible over \mathbb{Q} , by Eisenstein's criterion. Then f has exactly three real zeros with multiplicity 1 each, and the above lemma says that its Galois group is S_5 . To conclude, we need to show that the symmetric group S_n is not solvable if $n \geq 5$.

4.6 Solvability by ruler and compasses

The ancient Greek philosopher Plato believed that the only perfect figures were the straight line and the circle, and this belief had a great impact in ancient Greek geometry: it restricted the instruments available for performing geometrical constructions to ruler and compasses.

Many constructions can be done just by using ruler and compasses, but three famous constructions could not be performed:

- duplication of the cube: find a cube twice the volume of a given cube.
- trisection of the angle: find an angle $1/3$ the size of a given angle.
- quadrature of the circle: find a square of area equal to those of a given circle.

It is no wonder those problems remained unsolved (again, under these platonic constraints) since we will see, using our modern tools, that none of them are possible.

We start by formalizing the intuitive idea of a ruler and compass construction. Denote by P_0 the set of points in \mathbb{R}^2 .

- operation 1 (ruler): through any 2 points of P_0 , draw a straight line.
- operation 2 (compasses): draw a circle, whose center is a point of P_0 and whose radius is equal to the distance between some pairs of points in P_0 .

Definition 4.8. The points of intersection of any two distinct lines or circles, drawn using operations 1 and 2 are said to be **constructible** from P_0 if there exists a sequence r_1, \dots, r_n of points of \mathbb{R}^2 such that for each $i = 1, \dots, n$ the point r_i is constructible from the set $P_0 \cup \{r_1, \dots, r_{i-1}\}$, $P_i = P_{i-1} \cup \{r_i\}$.

We can now bring field theory into play. With each stage, we associate the subfield of \mathbb{R} generated by the coordinates of the points constructed. Denote by K_0 the subfield of \mathbb{R} generated by the x - and y -coordinates of the points in P_0 . If r_i has coordinates (x_i, y_i) , then inductively we define

$$K_i = K_{i-1}(x_i, y_i)$$

to get

$$K_0 \subseteq K_1 \subseteq \dots \subseteq K_n \subseteq \mathbb{R}.$$

Lemma 4.15. *With the above notation, x_i and y_i are zeros in K_i of quadratic polynomials over K_{i-1} .*

Proof. There are 3 cases to consider: line meets line, line meets circle and circle meets circle. We only give the proof of line meets circle.

Take 3 points $A = (p, q)$, $B = (r, s)$, $C = (t, u)$ in K_{i-1} , then draw a line between A and B , and a circle of center C with radius w . The equation of the line AB is

$$\frac{x-p}{r-p} = \frac{y-q}{s-q}$$

while the equation of the circle is

$$(x-t)^2 + (y-u)^2 = w^2.$$

Solving them yields

$$(x-t)^2 + \left(\frac{s-q}{r-p}(x-p) + q-u \right)^2 = w^2.$$

Now x , the first coordinate of the intersection point, is a zero of a quadratic polynomial over K_{i-1} . \square

We note that fields obtained by adjoining the zeroes of a quadratic polynomial are extensions of degree 2.

Theorem 4.16. *If $r = (x, y)$ is constructible from a subset $P_0 \in \mathbb{R}^2$, and if K_0 is the subfield of \mathbb{R} generated by the coordinates of the points of P_0 , then the degrees $[K_0(x) : K_0]$ and $[K_0(y) : K_0]$ are powers of 2.*

Proof. We have that

$$[K_{i-1}(x_i) : K_{i-1}] = 1 \text{ or } 2, \quad [K_{i-1}(y_i) : K_{i-1}] = 1 \text{ or } 2.$$

Using multiplication of degrees, we get

$$[K_{i-1}(x_i, y_i) : K_{i-1}] = [K_{i-1}(x_i, y_i) : K_{i-1}(x_i)][K_{i-1}(x_i) : K_{i-1}] = 1 \text{ or } 2 \text{ or } 4$$

with $K_i = K_{i-1}(x_i, y_i)$. Thus $[K_n : K_0]$ is a power of 2 implying that $[K_n : K_0(x)][K_0(x) : K_0]$ is a power of 2 from which we conclude that $[K_0(x) : K_0]$ is a power of 2, and similarly for y . \square

We are now ready to discuss the impossibility proofs.

Theorem 4.17. *The cube cannot be duplicated using ruler and compass constructions.*

Proof. Take a cube whose side is the unit interval, that is of volume 1. We have $P_0 = \{(0, 0), (1, 0)\}$ and $K_0 = \mathbb{Q}$. If we could duplicate the cube, then we can construct a point $(\alpha, 0)$ such that the volume α^3 is equal to 2, that is

$$\alpha^3 = 2.$$

Now $[\mathbb{Q}(\alpha) : \mathbb{Q}]$ is a power of 2, but α is a zero of $t^3 - 2$ which is irreducible (by Eisenstein) over \mathbb{Q} . This gives that

$$[\mathbb{Q}(\alpha) : \mathbb{Q}] = 3,$$

a contradiction to the fact that it should be a power of 2. \square

Theorem 4.18. *The angle $\pi/3$ cannot be trisected using ruler and compass constructions.*

Proof. Constructing an angle trisecting $\pi/3$ is equal to constructing the point $(\alpha, 0)$ given $(0, 0)$ and $(1, 0)$ where $\alpha = \cos(\pi/9)$. Knowing $\alpha = \cos(\pi/9)$, we can construct

$$\beta = 2 \cos(\pi/9).$$

Using that $\cos(3\theta) = 4 \cos^3(\theta) - 3 \cos(\theta)$ and $\cos(3\theta) = 1/2$ when $\theta = \pi/9$, we have

$$1 = 8 \cos^3(\theta) - 6 \cos(\theta) \Rightarrow \beta^3 - 3\beta - 1 = 0.$$

Now $f(t) = t^3 - 3t - 1$ is irreducible over \mathbb{Q} (apply Eisenstein on $f(t+1)$) thus

$$[\mathbb{Q}(\beta) : \mathbb{Q}] = 3$$

contradicting the fact that it should be a power of 2. \square

Theorem 4.19. *The circle cannot be squared using ruler and compass constructions.*

Proof. Without loss of generality, we assume that the circle is the unit circle centered at $(0, 0)$. Constructing a square with area π is equivalent to constructing a point $(\sqrt{\pi}, 0)$. Since the smallest field with 0 and 1 is \mathbb{Q} , the field obtained from adjoining $(\sqrt{\pi}, 0)$ is $\mathbb{Q}(\sqrt{\pi})$. Thus $[\mathbb{Q}(\sqrt{\pi}) : \mathbb{Q}]$ should be a power of 2, and in particular it should be algebraic, which is a contradiction (Lindeman's Theorem shows the transcendence of π , 1882). \square

The main definitions and results of this chapter are

- **(4.1).** Definitions of: Galois extension, Galois group, fixed field.
- **(4.2).** The fundamental theorem of Galois theory, Galois groups of intermediate fields.
- **(4.3).** Characterization of finite fields, their Galois group, their multiplicative group.
- **(4.4).** Definition of cyclotomic field, primitive root of unity, cyclotomic polynomial. The Galois group of a cyclotomic field.

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