FACULTY OF MATHEMATICS AND PHYSICS Charles University

## BACHELOR THESIS

Matouš Menčík

# The classical McKay correspondence 

Department of Algebra

Supervisor of the bachelor thesis: doc. RNDr. Jan Šťovíček, Ph.D.
Study programme: Mathematics
Study branch: General Mathematics

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Title: The classical McKay correspondence
Author: Matouš Menčík
Department: Department of Algebra
Supervisor: doc. RNDr. Jan Š̌̌ovíček, Ph.D., Department of Algebra
Abstract: The McKay correspondence is an interesting connection between many different areas of mathematics. The connecting element of the McKay correspondence is a special family of graphs called the Dynkin diagrams. In this thesis, we will study the classical McKay correspondence, which is an interesting connection between finite subgroups of $\mathrm{SL}(2, \mathrm{C})$ and Dynkin diagrams without oriented edges. Moreover, there are two ways how to get the Dynkin diagrams from the groups. In the first chapter of the thesis, we will provide a classification for both the finite subgroups and Dynkin diagrams. The second chapter uses the tools of the representation theory to construct the corresponding graph from the irreducible representations of the group. In the third part, we let the group act on the two-dimensional complex vector space. We then factor out this action to construct an algebraic variety with one singular point and find the Dynkin diagram in this singularity.

Keywords: McKay correspondence, McKay graphs, Dynkin diagrams

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## Introduction

In this thesis, we will study the classical McKay correspondence, which is a bijection between nontrivial finite subgroups of $S L(2, C)$ and the simply laced Dynkin diagrams. This correspondence has two parts, the algebraic and the geometric. Both describe a way how to construct a Dynkin diagram from a given subgroup of $S L(2, C)$, but the tools they are using are very different. We will firstly classify the finite subgroups of $S L(2, C)$ and the simply laced Dynkin diagrams and then cover both of the correspondences.
The structure of this thesis is inspired by Hemelsoet [2018], which covers similar topics. In this thesis, I've reached similar results, however, I tried to present my own proofs of the theorems.

## 1. Finite subgroups of $S L(2, \mathbb{C})$ and Dynkin diagrams

### 1.1 Classification of finite subgroups of $S L(2, \mathbb{C})$

Our first goal is to classify all the finite subgroups of $S L(2, \mathbb{C})$ up to conjugation. Fix any finite $G<S L(2, \mathbb{C})$ with more than two elements. Recall from linear algebra and group theory:

- Every matrix $g \in G$ satisfies $g^{|G|}=I$, where $I$ is the identity matrix.
- Over algebraically closed field, any matrix is conjugate to a matrix in Jordan canonical form.
- A Jordan block $B$ of dimension $d>1$ cannot satisfy an equality $B^{n}=I$ for $n>0$, therefore all matrices in $G$ must be diagonalizable.
- Two diagonalizable matrices commute if and only if they share a common eigenbasis. A proof of this statement can be found in Horn and Johnson 2012] (Theorem 1.3.12)
- Since $G<S L(2, \mathbb{C})$ and for diagonalizable matrices the determinant is the product of eigenvalues, the eigenvalues of any matrix from $G$ must be mutually inverse.

Now we can start the classification. Denote $\mathcal{A}$ the set of all maximal abelian subgroups of $G$ and a subgroup $Z<G$ depending on the parity of $G$ : $Z=\{I,-I\}$ for $|G|$ even and $Z=\{I\}$ for $|G|$ odd.

Proposition 1.1.1. The matrix $-I$ is in $G$ if and only if $|G|$ is even.
Proof. If $|G|$ is odd, then the order of any $g \in G$ is also odd, therefore $-I \notin G$. If $|G|$ is even, then it must contain an element of order 2. As shown above, this matrix must be conjugate to a matrix of the form $\operatorname{diag}\left(\lambda, \lambda^{-1}\right)$. Since order of this matrix is 2 , the eigenvalues must satisfy the equation $\lambda^{2}=1$, therefore $\lambda \in\{1,-1\}$. $\lambda=1$ gives us a matrix $I$ of order $1, \lambda=-1$ gives us $-I$. These matrices commute with any other matrix, therefore conjugation doesn't change them and we see that $-I$ is the only possible matrix of order 2 , which can be in $G$.

Proposition 1.1.2. Let $A_{1}, A_{2} \in \mathcal{A}, A_{1} \neq A_{2}$. Then $A_{1} \cap A_{2}=Z$.
Proof. Since $Z$ is a subgroup of the center $Z(S L(2, \mathbb{C}))$, it is also in the center $Z(G)$. If we take a group generated by the elements of $A_{1}$ and elements of $Z$, its generators commute, therefore the group is abelian. From maximality of $A_{1}$ this means that $Z \leq A_{1}$. Similarly $Z \leq A_{2}$, therefore $Z \leq A_{1} \cap A_{2}$. Now suppose $b \in\left(A_{1} \cap A_{2}\right) \backslash Z$. The only matrices in $G$, which have an eigenspace of dimension $d>1$ are in $Z$, therefore $b$ has only one eigenbasis $B$ up to permuting and scaling the vectors. From which it follows that for any matrix in $G$ commuting with $b, B$ is an eigenbasis. Denote $A=\{g \in G \mid B$ is an eigenbasis of $g\}$. This set is closed
under multiplication, inverses and contains $I$, therefore a subgroup of $G$. From the property of sharing an eigenbasis, $A$ is abelian. Therefore $A_{1} \leq A \geq A_{2}$ and from the maximality of $A_{1}, A_{2}$ follows $A_{1}=A=A_{2}$.

Corollary 1.1.3. $|G|+(|\mathcal{A}|-1)|Z|=\sum_{A \in \mathcal{A}}|A|$
Proof. Every element of $G \backslash Z$ is in some $A \in \mathcal{A}$, from the Proposition in exactly one, therefore $|G|-|Z|=\sum_{A \in \mathcal{A}}(|A|-|Z|)=\sum_{A \in \mathcal{A}}|A|-|\mathcal{A}| \cdot|Z|$.
Proposition 1.1.4. For every $A \in \mathcal{A}$ there exists a matrix $M \in G L(2, \mathbb{C})$ and a natural number $n$, such that $A=\left\{\left.M \cdot \operatorname{diag}\left(e^{\frac{2 k \pi i}{n}}, e^{-\frac{2 k \pi i}{n}}\right) \cdot M^{-1} \right\rvert\, k \in\{0, \ldots, n-1\}\right\}$.
Proof. The set $Z$ is not a maximal abelian subgroup of $G$ (the group generated by elements of $Z$ and any other element is abelian), therefore for all $A \in \mathcal{A}$ it holds $A \backslash Z \neq \emptyset$. According to the proof of previous proposition, if we choose any $b \in a \backslash Z$ with eigenbasis $B$, then $A=\{g \in G \mid B$ is an eigenbasis of $g\}$. Therefore if we define $M$ as the matrix, whose columns are the vectors of $B, A$ will be of the form $\left\{M \cdot \operatorname{diag}\left(\lambda_{i}, \lambda_{i}^{-1}\right) \cdot M^{-1} \mid i \in\{1, \ldots,|A|\}\right\}$ for some $\lambda_{i} \in \mathbb{C}$. Since the order of every $a \in A$ is a divisor of $|G|$, only possible $\lambda_{i}$ are those of the form $e^{\frac{2 k \pi i}{\mid G i}}$ for $k \in\{0, \ldots,|G|-1\}$. Therefore $A$ is a subgroup of a cyclic group $\left\{\left.M \cdot \operatorname{diag}\left(e^{\frac{2 k \pi i}{|G|}}, e^{-\frac{2 k \pi i}{|G|}}\right) \cdot M^{-1} \right\rvert\, k \in\{0, \ldots,|G|-1\}\right\}$, from which it follows it is of the form we claimed.

Proposition 1.1.5. If $A \in \mathcal{A}$ and $g$ is in the normalizer $N(A)$, then either gag $^{-1}=a$ for every $a \in N(A) \backslash A$, or gag $g^{-1}=a^{-1}$ for every $a \in N(A) \backslash A$.

Proof. Since $A$ is cyclic, any homomorphism from $A$ is given by its value on a generator of $A$. The map $a \mapsto g a g^{-1}$ is a homomorphism and conjugation doesn't change the eigenvalues of a matrix, therefore only possible values of $\mathrm{gag}^{-1}$ are $a$ and $a^{-1}$.

Proposition 1.1.6. For $A \in \mathcal{A}$, either $N(A)=A$ or $|N(A)|=2|A|$.
Proof. Define a homomorphism $N(A) \rightarrow \operatorname{Aut}(A), g \mapsto \varphi_{g}, \varphi_{g}(a)=g a g^{-1}$. The image of this homomorphism has one or two elements, the kernel consists of those elements, which commute with all elements from $A$, from maximality of $A$ this kernel is $A$.

Theorem 1.1.7. For odd $|G|, G$ is cyclic. For even $|G|, G$ is cyclic or satisfy the equality $|\mathcal{A}|=\frac{|G|}{4}+1$.
Proof. When we consider the action of conjugation on $\mathcal{A}$, i.e. $\varphi: G \rightarrow S(\mathcal{A})$, $\varphi_{g}(A)=g A g^{-1}$ and denote the orbit of $A$ as $\operatorname{Conj}(A)$, then $|\operatorname{Conj}(A)|=\frac{|G|}{|N(A)|}$. Therefore if we denote $l_{A}=\frac{|N(A)|}{|A|}$, we get an equality $|A|=\frac{|G|}{l_{A}|C o n j(A)|}$. From Corollary 1.1.3 then follows

$$
1+\frac{(|\mathcal{A}|-1)|Z|}{|G|}=\sum_{A \in \mathcal{A}} \frac{1}{l_{A}|\operatorname{Conj}(A)|} .
$$

Conjugate groups have conjugate normalizers, therefore if we define $\mathcal{C}$ a set of orbits $\{\operatorname{Conj}(A) \mid A \in \mathcal{A}, C=\operatorname{Conj}(A)\}$, we can rewrite right-hand side of the equation and get

$$
1+\frac{(|\mathcal{A}|-1)|Z|}{|G|}=\sum_{A \in \mathcal{A}} \frac{1}{l_{A}|\operatorname{Conj}(A)|}=\sum_{C \in \mathcal{C}} \sum_{A \in C} \frac{1}{l_{A}|\operatorname{Conj}(A)|}=\sum_{C \in \mathcal{C}} \frac{1}{l_{C}},
$$

where $l_{C}$ is defined as $l_{A}$ of any $A \in C$. But since all $l_{A}$ are 1 or 2 , the right-hand side is now some integral multiple of $\frac{1}{2}$. Therefore left-hand side must be also, so $|G|$ divides $2(|\mathcal{A}|-1)|Z|$. For $|G|$ odd, all $A \in \mathcal{A}$ must also have odd order, therefore at least 3. For $|G|$ even, they all contain $-I$, therefore they must have even order, therefore at least 4 . In both cases, $A \backslash Z$ always contains at least 2 elements, which are not in any other $A_{i} \backslash Z$, therefore $|G|>2|\mathcal{A}|$. This means that $2(|\mathcal{A}|-1)|Z|<|G| \cdot|Z|$. For $|G|$ odd this means $2(|\mathcal{A}|-1)|Z|=0|G|$, therefore $|\mathcal{A}|=1$ and $\mathcal{A}=\{G\}$. Since all $A \in \mathcal{A}$ are cyclic, $G$ is cyclic. For even $|G|$, we can either have $2(|\mathcal{A}|-1)|Z|=0|G|$ and for the same reason is $G$ cyclic, or $2(|\mathcal{A}|-1)|Z|=1|G|$, from which follows the statement.

Remark 1.1.8. From the proof it also follows that in the case of non-cyclic groups, $\sum_{C \in \mathcal{C}} \frac{1}{l_{C}}=\frac{3}{2}$, so $\mathcal{C}$ has either two elements with one having $l_{C_{1}}=1$ and the other $l_{C_{1}}=2$, or three elements with $l_{C_{1}}=l_{C_{2}}=l_{C_{3}}=2$

From now on, assume $G$ is not cyclic (and therefore $|G|$ is even and $|Z|=2$ ).
Proposition 1.1.9. If $A_{1}, A_{2}$ or $A_{1}, A_{2}, A_{3}$, are representatives of different elements of $\mathcal{C}$, then $\frac{1}{|G|}+\frac{1}{4}=\sum_{i} \frac{1}{\left|A_{i}\right| A_{i} \mid}$

Proof. Follows from the equality $\frac{|G|}{4}+1=\mathcal{A}=\sum_{i}\left|\operatorname{Conj}\left(A_{i}\right)\right|$ when divided by $|G|$ and using the equality $|A|=\frac{|G|}{l_{A}|\operatorname{Conj}(A)|}$.

Proposition 1.1.10. If $|\mathcal{C}|=3$, the only values of $\left|A_{i}\right|$ and $|G|$ satisfying the equality $\frac{1}{G}+\frac{1}{4}=\sum_{i} \frac{1}{l_{A_{i}}\left|A_{i}\right|}$ are (up to permutation of $A_{i}$ ):

| Type: | $\left\|A_{1}\right\|$ | $\left\|A_{2}\right\|$ | $\left\|A_{3}\right\|$ | $\|G\|$ |
| :---: | :---: | :---: | :---: | :---: |
| $I$ | 4 | 4 | $2 k$ | $4 k$ |
| II | 4 | 6 | 6 | 24 |
| III | 4 | 6 | 8 | 48 |
| IV | 4 | 6 | 10 | 120 |

If $|\mathcal{C}|=2$, the only possible values satisfying the equation (where $l_{A_{1}}=1, l_{A_{2}}=2$ ) are:

| Type: | $\left\|A_{1}\right\|$ | $\left\|A_{2}\right\|$ | $\|G\|$ |
| :---: | :---: | :---: | :---: |
| $V$ | 4 | $2 k$ | $4 k$ |
| $V I$ | 6 | 4 | 24 |

Proof. Every $A \in \mathcal{A}$ contains the element $-I$ of order 2 , therefore $|A|$ is even. $\{I,-I\}$ is not maximal, therefore $|A| \geq 4$. If $|\mathcal{C}|=3$, we may WLOG assume $\left|A_{1}\right| \leq\left|A_{2}\right| \leq\left|A_{3}\right|$. Then $\left|A_{1}\right|=4$, since $6 \leq\left|A_{1}\right| \leq\left|A_{2}\right| \leq\left|A_{3}\right|$ implies $\sum_{i} \frac{1}{2\left|A_{i}\right|} \leq \frac{1}{4}<\frac{1}{4}+\frac{1}{|G|}$. If $\left|A_{2}\right|=4$, any value $\left|A_{3}\right|=2 k$ satisfies the equation with $|G|=4 k$. If $\left|A_{2}\right|=6$, the only possible values for $\left|A_{3}\right|$ are 6,8 and 10 , otherwise again $\sum_{i} \frac{1}{2\left|A_{i}\right|} \leq \frac{1}{4}<\frac{1}{4}+\frac{1}{|G|}$. For the same reason $8 \leq\left|A_{2}\right|$ does not lead to a solution. If $|\mathcal{C}|=2$ and $l_{A_{1}}=1, l_{A_{2}}=2$, then $\left|A_{1}\right|=x,\left|A_{2}\right|=y$ is a solution of the equation if and only if $\left|A_{1}\right|=\left|A_{2}\right|=x,\left|A_{3}\right|=y$ is a solution of the case $|\mathcal{C}|=3$.

Proposition 1.1.11. The group $G$ has $2+\sum_{i} \frac{\left|A_{i}\right|-2}{l_{A_{i}}}$ conjugacy classes.
Proof. Besides $\pm I$, there are $\left|A_{i}\right|-2$ elements in each $A_{i}$, therefore there exist $\left(\left|A_{i}\right|-2\right)\left|\operatorname{Conj}\left(A_{i}\right)\right|=\left(\left|A_{i}\right|-2\right) \frac{|G|}{l_{A_{i}}\left|A_{i}\right|}$ such elements in $\cup \operatorname{Conj}\left(A_{i}\right)$. Each of
these elements commutes only with its maximal abelian subgroup, therefore has conjugacy class of size $\frac{|G|}{\left|A_{i}\right|}$. Therefore $\cup \operatorname{Conj}\left(A_{i}\right) \backslash\{ \pm I\}$ splits into $\frac{\left|A_{i}\right|-2}{l_{A_{i}}}$ conjugacy classes. If we sum over $i$ and add the conjugacy classes $\{I\}$ and $\{-I\}$, we get the expression $2+\sum_{i} \frac{\left|A_{i}\right|-2}{l_{A_{i}}}$.

Proposition 1.1.12. For every $n \in \mathbb{N}, S L(2, \mathbb{C})$ has exactly one cyclic subgroup of order $n$ up to conjugation.
Proof. The group $\left\{\left.\operatorname{diag}\left(e^{\frac{2 k \pi i}{n}}, e^{-\frac{2 k \pi i}{n}}\right) \right\rvert\, k \in\{0, \ldots, n-1\}\right\}$ satisfies the properties, therefore existence is proved. From Proposition 1.1.4 it follows that any two such groups are conjugate.

Proposition 1.1.13. The subgroup of type I exists if and only if $k$ is even. The subgroup of type $V$ exists if and only if $k$ is odd and greater than 1 . In both cases, the group is unique up to conjugation.

Proof. Let $G$ be group of the type $I$ or $V$. According to Proposition 1.1.4, we can find a matrix $S$ with eigenvalues $e^{\frac{i \pi}{k}}, e^{-\frac{i \pi}{k}}$ in the $2 k$-element cyclic subgroup and a matrix $T$ with eigenvalues $e^{\frac{i \pi}{2}}, e^{-\frac{i \pi}{2}}$ in one of the 4 -element maximal cyclic subgroups. Since these matrices do not commute and $|\operatorname{Conj}(\langle S\rangle)|=\frac{|G|}{l_{\langle S\rangle}|\langle S\rangle|}=1$ (in other words, $\langle S\rangle$ is normal), it must hold $T S T^{-1}=S^{-1}$. Let $B=\left(b_{1}, b_{2}\right)$ be a basis, such that

$$
[S]_{B}=\left(\begin{array}{cc}
e^{\frac{i \pi}{k}} & 0 \\
0 & e^{-\frac{i \pi}{k}}
\end{array}\right)
$$

If we denote the elements of $[T]_{B}$ as

$$
[T]_{B}=\left(\begin{array}{ll}
q & r \\
s & t
\end{array}\right)
$$

we can compute its inverse

$$
[T]_{B}^{-1}=\frac{1}{q t-r s}\left(\begin{array}{cc}
t & -r \\
-s & q
\end{array}\right)
$$

where from $T \in S L(2, \mathbb{C})$ it follows that $q t-r s=\operatorname{det}(T)=1$, and get a matrix equality

$$
\left(\begin{array}{ll}
q t e^{\frac{i \pi}{k}}-r s e^{-\frac{i \pi}{k}} & q r\left(e^{-\frac{i \pi}{k}}-e^{\frac{i \pi}{k}}\right) \\
\operatorname{st}\left(e^{\frac{i \pi}{k}}-e^{-\frac{i \pi}{k}}\right) & q t e^{-\frac{i \pi}{k}}-r s e^{\frac{i \pi}{k}}
\end{array}\right)=\left[T S T^{-1}\right]_{B}=\left[S^{-1}\right]_{B}=\left(\begin{array}{cc}
e^{-\frac{i \pi}{k}} & 0 \\
0 & e^{\frac{i \pi}{k}}
\end{array}\right) .
$$

From the equality of top right corners we learn that either $q=0$ or $r=0$, from the equality of bottom left corners we learn that either $s=0$ or $t=0$. Since $\operatorname{det}(T)=q t-r s=1$, this gives us only two possibilities, $[T]_{B}=\left(\begin{array}{cc}\lambda & 0 \\ 0 & \lambda^{-1}\end{array}\right)$ or $[T]_{B}=\left(\begin{array}{cc}0 & -\lambda \\ \lambda^{-1} & 0\end{array}\right)$ for some $\lambda \in \mathbb{C}^{*}$, out of which only the second one gives us the equality of the diagonal elements. If we now define a basis $B^{\prime}=\left(\lambda b_{1}, b_{2}\right)$, we get

$$
[S]_{B^{\prime}}=\left(\begin{array}{cc}
e^{\frac{i \pi}{k}} & 0 \\
0 & e^{-\frac{i \pi}{k}}
\end{array}\right),[T]_{B^{\prime}}=\left(\begin{array}{cc}
0 & -1 \\
1 & 0
\end{array}\right) .
$$

Since $G$ must be generated by $S$ and $T$ (the subgroup $\langle S, T\rangle$ contains an element of order $2 k$ and is not abelian, there is no such subgroup of $S L(2, \mathbb{C})$ with order less than $4 k$ ), this proves the uniqueness. To show the existence, let

$$
S=\left(\begin{array}{cc}
e^{\frac{i \pi}{k}} & 0 \\
0 & e^{-\frac{i \pi}{k}}
\end{array}\right), T=\left(\begin{array}{cc}
0 & -1 \\
1 & 0
\end{array}\right)
$$

and $G=\langle S, T\rangle$. Since $T\langle S\rangle T^{-1}=\langle S\rangle,\langle S\rangle$ is a normal subgroup with $2 k$ elements. Since $T \notin\langle S\rangle$, but $T^{2}=-I \in\langle S\rangle,\left|\frac{\langle S, T\rangle}{\langle S\rangle}\right|=2$. Therefore $\langle S, T\rangle$ has $4 k$ elements. It is not abelian and contains an element of order $2 k$, therefore it is either type $I$ or type $V$. If $k$ is odd, it can't be of the type $I$, because the number of sets conjugate to $A_{1}$ would have to be $\frac{|G|}{\left|A_{1}\right| \cdot l_{A_{1}}}=\frac{k}{2}$, which is not an integer. If $k$ is even, then $S^{k / 2} T S^{-k / 2}=S^{k / 2} S^{k / 2} T=-I T=T^{3}=T^{-1}$, from which it follows that $l_{\langle T\rangle}=2$, therefore $G$ is a group of the type $I$.

Proposition 1.1.14. A group of type II does not exist. A group of type VI exists and is unique up to conjugation.

Proof. Let $G$ be a group of the type $I I$ or $V I$. In both cases, there are four maximal abelian subgroup of order 6 , therefore eight matrices with eigenvalues $e^{\frac{\pi i}{3}}$ and $e^{-\frac{\pi i}{3}}$. Denote them $M_{1}, M_{2}, M_{3}, M_{4}, M_{1}^{-1}, M_{2}^{-1}, M_{3}^{-1}$ and $M_{4}^{-1}$. The group $\left\langle M_{1}\right\rangle$ acts on $\left\{M_{2}, M_{3}, M_{4}, M_{2}^{-1}, M_{3}^{-1}, M_{4}^{-1}\right\}$ by conjugation. Since an element of the set commute with an element of the group if and only if the element of the group is $\pm I$, the stabilizer of every element is $\{ \pm I\}$ and there are two orbits of size 3. Suppose $M_{i}^{-1}=M_{1}^{n} M_{i} M_{1}^{-n}$. Then for any $M_{j}=M_{1}^{m} M_{i} M_{1}^{-m}$ it holds $M_{j}^{-1}=M_{1}^{m} M_{i}^{-1} M_{1}^{-m}=M_{1}^{n+m} M_{i} M_{1}^{-n-m}=M_{1}^{n} M_{j} M_{1}^{-n}$, therefore if one element in the orbit has its inverse in the same orbit, than this holds for all elements of the orbit. Therefore this can't happen, since the orbits have odd number of elements. Therefore $\left|\operatorname{Conj}\left(\left\langle M_{2}\right\rangle\right)\right| \geq 3$ and $G$ is of type VI. WLOG, we can therefore assume $M_{1}, M_{2}, M_{3}$ and $M_{4}$ lie in the same conjugacy class, and $M_{1}$ act on the rest by conjugation as permutation $\left(M_{2} M_{3} M_{4}\right)$.

Since $G$ is of the type $V I$, none of the $M_{i}$ is conjugate to its inverse, therefore orbit of $M_{1}$ under the action of conjugating by $\left\langle M_{2}\right\rangle$ is $\left\{M_{1}, M_{3}, M_{4}\right\}$. Let $M_{2}$ act on this set as a permutation $\left(M_{1} M_{i} M_{j}\right)$ for $\{i, j\}=\{3,4\}$. In the case $i=3, j=4$ we reach a contradiction, since conjugating by $M_{3}=M_{1} M_{2} M_{1}^{-1}$ would give us a permutation $\left(M_{2} M_{3} M_{4}\right) \circ\left(M_{1} M_{3} M_{4}\right) \circ\left(M_{2} M_{3} M_{4}\right)^{-1}=$ $\left(M_{1} M_{4} M_{2}\right)$, but $M_{3}=M_{2} M_{1} M_{2}^{-1}$ would give us $\left(M_{1} M_{3} M_{4}\right) \circ\left(M_{2} M_{3} M_{4}\right) \circ$ $\left(M_{1} M_{3} M_{4}\right)^{-1}=\left(M_{1} M_{2} M_{4}\right)$. Therefore $i=4, j=3$. From this follows that the matrix $M_{1} M_{2}^{-1}$ acts as a permutation $\left(M_{2} M_{3} M_{4}\right) \circ\left(M_{1} M_{4} M_{3}\right)^{-1}=$ $\left(M_{1} M_{4}\right)\left(M_{2} M_{3}\right)$. Therefore $M_{1} M_{2}^{-1}$ is a group element of order 4. Let $B=$ $\left(b_{1}, b_{2}\right)$ be a basis, such that

$$
\left[M_{1}\right]_{B}=\left(\begin{array}{cc}
e^{\frac{\pi i}{3}} & 0 \\
0 & e^{-\frac{\pi i}{3}}
\end{array}\right)
$$

Suppose

$$
\left[M_{2}^{-1}\right]_{B}=\left(\begin{array}{ll}
q & r \\
s & t
\end{array}\right),
$$

therefore

$$
\left[M_{1} M_{2}^{-1}\right]_{B}=\left(\begin{array}{ll}
q e^{\frac{\pi i}{3}} & r e^{-\frac{\pi i}{3}} \\
s e^{\frac{\pi i}{3}} & t e^{-\frac{\pi i}{3}}
\end{array}\right) .
$$

The matrix $M_{2}^{-1}$ is in $S L(2, \mathbb{C})$, therefore $q t-r s=1$. It has eigenvalues $e^{\frac{\pi i}{3}}$ and $e^{-\frac{\pi i}{3}}$, therefore $q+t=\operatorname{tr}\left(\left[M_{2}^{-1}\right]_{B}\right)=1$. The matrix $M_{1} M_{2}^{-1}$ has eigenvalues $\pm i$, therefore $q e^{\frac{\pi i}{3}}+t e^{-\frac{\pi i}{3}}=\operatorname{tr}\left(\left[M_{1} M_{2}^{-1}\right]_{B}\right)=0$. From the last two equalities follows $0=q e^{\frac{\pi i}{3}}+(1-q) e^{-\frac{\pi i}{3}}=\frac{1-\sqrt{3} i}{2}+\sqrt{3} i q, q=\frac{3+\sqrt{3} i}{6}, t=\frac{3-\sqrt{3} i}{6}$. Therefore

$$
\left[M_{2}^{-1}\right]_{B}=\left(\begin{array}{cc}
\frac{3+\sqrt{3} i}{6} & r \\
-\frac{2}{3 r} & \frac{3-\sqrt{3} i}{6}
\end{array}\right)
$$

for some $r \in \mathbb{C}^{*}$. Again, we can change the basis to $B^{\prime}=\left(r b_{1}, b_{2}\right)$ and then

$$
\begin{gathered}
{\left[M_{1}\right]_{B^{\prime}}=\left(\begin{array}{cc}
e^{\frac{\pi i}{3}} & 0 \\
0 & e^{-\frac{\pi i}{3}}
\end{array}\right)} \\
{\left[M_{2}^{-1}\right]_{B^{\prime}}=\left(\begin{array}{cc}
\frac{3+\sqrt{3} i}{6} & 1 \\
-\frac{2}{3} & \frac{3-\sqrt{3} i}{6}
\end{array}\right) .}
\end{gathered}
$$

The group $\left\langle M_{1}, M_{2}\right\rangle$ must be the whole $G$, since there is no subgroup of $S L(2, \mathbb{C})$ with order less then 24 and two noncommuting elements of order 6 . This proves the uniqueness. To prove the existence, let

$$
\begin{gathered}
M_{1}=\left(\begin{array}{cc}
e^{\frac{\pi i}{3}} & 0 \\
0 & e^{-\frac{\pi i}{3}}
\end{array}\right) \\
M_{2}=\left(\begin{array}{cc}
\frac{3-\sqrt{3} i}{6} & -1 \\
\frac{2}{3} & \frac{3+\sqrt{3} i}{6}
\end{array}\right) \\
M_{3}=\left(\begin{array}{cc}
\frac{3-\sqrt{3} i}{6} & \frac{1-\sqrt{3} i}{2} \\
\frac{-1-\sqrt{3} i}{3} & \frac{3+\sqrt{3} i}{6}
\end{array}\right) \\
M_{4}=\left(\begin{array}{ll}
\frac{3-\sqrt{3} i}{6} & \frac{1+\sqrt{3} i}{2} \\
\frac{-1+\sqrt{3} i}{3} & \frac{3+\sqrt{3} i}{6}
\end{array}\right)
\end{gathered}
$$

By direct computation, one can show that this set of matrices is closed under under conjugation by $M_{1}$ and $M_{2}$, which act on it as the already mentioned permutations $\left(M_{2} M_{3} M_{4}\right)$ and ( $M_{1} M_{2} M_{4}$ ) respectively. Therefore the action of conjugation of $M_{1}, M_{2}, M_{3}$ and $M_{4}$ by $\left\langle M_{1}, M_{2}\right\rangle$ can be viewed as a group homomorphism $\left\langle M_{1}, M_{2}\right\rangle \rightarrow A_{4}$. This homomorphism has kernel $\{ \pm I\}$, since there is no other element commuting with all of these matrices. Therefore $\left\langle M_{1}, M_{2}\right\rangle$ is a group of order at most 24 , since it contains two noncommuting elements of order 6 , it is of the type $V I$.

Proposition 1.1.15. A group of type III does exist and is unique up to conjugation.

Proof. Let $G$ be a group of the type III. There it contains 6 matrices with eigenvalues $e^{\frac{i \pi}{4}}$ and $e^{\frac{i \pi}{4}}$, denote them $M_{1}, M_{2}, M_{3}, M_{1}^{-1}, M_{2}^{-1}$ and $M_{3}^{-1}$. The set
$\left\{M_{2}, M_{3}, M_{2}^{-1}, M_{3}^{-1}\right\}$ is closed under conjugation by $\left\langle M_{1}\right\rangle$. Since the only elements from $\left\langle M_{1}\right\rangle$, which commute with some element from $\left\{M_{2}, M_{3}, M_{2}^{-1}, M_{3}^{-1}\right\}$, are $\pm I$, the kernel of this action is $\{ \pm I\}$ and the action is transitive. The conjugation of $M_{2}$ by $M_{1}$ cannot be $M_{2}^{-1}$, otherwise the conjugation of $M_{2}^{-1}$ by $M_{1}$ would be $M_{2}$ and the orbit would have only these two elements. Therefore we can WLOG assume $M_{1}$ acts on $\left\{M_{2}, M_{3}, M_{2}^{-1}, M_{3}^{-1}\right\}$ as the permutation $\left(M_{2} M_{3} M_{2}^{-1} M_{3}^{-1}\right)$. For the same reason, the matrix $M_{2}$ acts on $\left\{M_{1}, M_{3}, M_{1}^{-1}, M_{3}^{-1}\right\}$ as a permutation $\left(M_{1}^{ \pm 1} M_{3} M_{1}^{\mp 1} M_{3}^{-1}\right)$. The matrix $M_{1} M_{2}^{2}$ therefore acts on the set of all eight matrices as $\left(M_{2} M_{3} M_{2}^{-1} M_{3}^{-1}\right) \circ\left(M_{1}^{ \pm 1} M_{3} M_{1}^{\mp 1} M_{3}^{-1}\right)^{2}$, therefore as a permutation $\left(M_{1} M_{1}^{-1}\right)\left(M_{2} M_{3}\right)\left(M_{2}^{-1} M_{3}^{-1}\right)$. From this follows $M_{1} M_{2}^{2}$ is an element of order 4 . Let $B=\left(b_{1}, b_{2}\right)$ be a basis, such that

$$
\left[M_{2}\right]_{B}=\left(\begin{array}{cc}
e^{\frac{\pi i}{4}} & 0 \\
0 & e^{-\frac{\pi i}{4}}
\end{array}\right) .
$$

Suppose

$$
\left[M_{1}\right]_{B}=\left(\begin{array}{ll}
q & r \\
s & t
\end{array}\right)
$$

therefore

$$
\left[M_{1} M_{2}^{2}\right]_{B}=\left(\begin{array}{ll}
i q & -i r \\
i s & -i t
\end{array}\right)
$$

Since $M_{1}$ has eigenvalues $e^{\frac{i \pi}{4}}$ and $e^{\frac{i \pi}{4}}$, its trace is $\sqrt{2}$. Since $M_{1} M_{2}^{2}$ has order 4, it has eigenvalues $\pm i$ and its trace is zero. Therefore $q+t=\sqrt{2}, i q-i t=0$, which has solution $q=t=\frac{\sqrt{2}}{2}$. Since $M_{1} \in S L(2, \mathbb{C})$, it must be of the form

$$
\left[M_{1}\right]_{B}=\left(\begin{array}{cc}
\frac{\sqrt{2}}{2} & r \\
-\frac{1}{2 r} & \frac{\sqrt{2}}{2}
\end{array}\right) .
$$

Again, we change the basis to $B^{\prime}=\left(\sqrt{2} r b_{1}, b_{2}\right)$ to get

$$
\begin{aligned}
{\left[M_{2}\right]_{B^{\prime}} } & =\left(\begin{array}{cc}
e^{\frac{\pi i}{4}} & 0 \\
0 & e^{-\frac{\pi i}{4}}
\end{array}\right) \\
{\left[M_{1}\right]_{B^{\prime}} } & =\left(\begin{array}{cc}
\frac{\sqrt{2}}{2} & \frac{\sqrt{2}}{2} \\
-\frac{\sqrt{2}}{2} & \frac{\sqrt{2}}{2}
\end{array}\right) .
\end{aligned}
$$

Since $M_{1}$ and $M_{2}$ does not commute and both have order 8, the group $\left\langle M_{1}, M_{2}\right\rangle$ must be the whole $G$, which proves the uniqueness. To prove the existence, let

$$
\begin{aligned}
& M_{2}=\left(\begin{array}{cc}
e^{\frac{\pi i}{4}} & 0 \\
0 & e^{-\frac{\pi i}{4}}
\end{array}\right) \\
& M_{1}=\left(\begin{array}{cc}
\frac{\sqrt{2}}{2} & \frac{\sqrt{2}}{2} \\
-\frac{\sqrt{2}}{2} & \frac{\sqrt{2}}{2}
\end{array}\right) .
\end{aligned}
$$

Since the set of 32 vectors

$$
\left\{\omega v \left\lvert\, v \in\left\{\binom{1}{0},\binom{0}{1},\binom{\frac{\sqrt{2}}{2}}{\frac{\sqrt{2}}{2}},\binom{\frac{\sqrt{2}}{2}}{-\frac{\sqrt{2}}{2}}\right\}\right., \omega \in\left\{ \pm 1, \pm i, \pm \frac{\sqrt{2}}{2} \pm \frac{\sqrt{2}}{2} i\right\}\right\}
$$

is closed under multiplying by $M_{1}$ and $M_{2}$, it is closed under multiplying by all matrices from $\left\langle M_{1}, M_{2}\right\rangle$. These vectors generate the space $\mathbb{C}^{2}$, therefore a matrix is uniquely determined by its action on these vectors and the $\left\langle M_{1}, M_{2}\right\rangle$ can be embedded to $S_{32}$. Therefore it is finite and since it contains two noncommuting elements of order 8 , it is of the type III.

Proposition 1.1.16. A group of type IV exists and is unique up to conjugation.
Proof. Let $G$ be a group of type $I V$. It contains a matrix $M_{1}$ with eigenvalues $e^{\frac{i \pi}{5}}$ and $e^{-\frac{i \pi}{5}}$. Let $B=\left(b_{1}, b_{2}\right)$ be a basis, such that

$$
\left[M_{1}\right]_{B}=\left(\begin{array}{cc}
e^{\frac{\pi i}{5}} & 0 \\
0 & e^{-\frac{\pi i}{5}}
\end{array}\right) .
$$

Let $M_{2}$ be another matrix with the same eigenvalues, which generates a different maximal abelian subgroup. Suppose

$$
\left[M_{2}\right]_{B}=\left(\begin{array}{ll}
q & r \\
s & t
\end{array}\right) .
$$

Since $M_{2}$ has eigenvalues $e^{\frac{i \pi}{5}}$ and $e^{-\frac{i \pi}{5}}$, it has trace $q+t=2 \cos \left(\frac{\pi}{5}\right)=\frac{1+\sqrt{5}}{4}$. Since $M_{1} M_{2}$ has order $3,5,6,10$ or 4 , its trace $e^{\frac{\pi i}{5}} q+e^{-\frac{\pi i}{5}} t$ can be $2 \cos (2 \pi \alpha)$ for $\alpha \in\left\{\frac{1}{10}, \frac{2}{10}, \frac{3}{10}, \frac{4}{10}, \frac{1}{6}, \frac{2}{6}, \frac{1}{4}\right\}$. The same holds for the matrix $M_{1}^{-1} M_{2}$ with the trace $e^{-\frac{\pi i}{5}} q+e^{\frac{\pi i}{5}} t$. If we add these equations, we get $2 \cos \left(2 \pi \alpha_{1}\right)+2 \cos \left(2 \pi \alpha_{2}\right)=$ $2 \cos \left(\frac{\pi}{5}\right) q+2 \cos \left(\frac{\pi}{5}\right) t=4 \cos ^{2}\left(\frac{\pi}{5}\right)$. The only solution with possible $\alpha_{1}, \alpha_{2}$ is $\alpha_{1}=\frac{1}{10}$, $\alpha_{2}=\frac{1}{6}$ or $\alpha_{1}=\frac{1}{6}, \alpha_{2}=\frac{1}{10}$. We can WLOG assume $M_{1} M_{2}$ has eigenvalues $e^{\frac{i \pi}{5}}$ and $e^{-\frac{i \pi}{5}}$, otherwise we can redefine $M_{2}$ to be $M_{2}^{-1}$ and then $M_{1} M_{2}^{\text {new }}=\left(M_{2}^{\text {old }} M_{1}^{-1}\right)^{-1}$, which is a matrix conjugated to $M_{1}^{-1} M_{2}^{\text {old }}$ and have the same eigenvalues. We can then compute $q$ and $t$ from $q+t=\frac{1+\sqrt{5}}{4}$ and $e^{\frac{\pi i}{5}} q+e^{-\frac{\pi i}{5}} t=\frac{1+\sqrt{5}}{4}$ and get $q=\frac{1+\sqrt{5}}{4} \frac{1-e^{\frac{\pi i}{2 i}}}{1-e^{\frac{2 \pi i}{5}}}=\cdot \frac{1+\sqrt{5}}{4\left(1+e^{\frac{\pi i}{5}}\right)}, t=\frac{(1+\sqrt{5}) e^{\frac{\pi i}{2}}}{4\left(1+e^{\frac{\pi i}{5}}\right)}$. The same way as in the previous cases, we can change the basis, such that both matrices $M_{1}, M_{2}$ have with respect to that basis some fixed form. Since there is no other subgroup of $S L(2, \mathbb{C})$ with two noncommuting matrices of order 10, this shows uniqueness. Since elements of this matrices are in very complicated form, we will not prove the existence this way. The proof of the can be found in Lamotke 1986 (page 35).

Definition 1.1.17. The cyclic group of order $n$ is denoted by $C_{n}$. The group of type $I$ or $V$ of order $4 n$ is called the binary dihedral group and is denoted by $B D_{4 n}$. The group of type VI is called the binary tetrahedral group and is denoted by $B T$. The group of type III is called the binary octahedral group and denoted by $B O$. The group of type IV is called the binary icosahedral group and denoted by $B I$.

The names of groups $B T, B O$ and $B I$ are derived from the names of the platonic solids, because the quotient group $G /\{ \pm I\}$ is isomorphic to the rotation group of the corresponding solid. More about this also in Lamotke 1986.

### 1.2 Dynkin diagrams

Definition 1.2.1. A multigraph is a triple $(V, E, f)$, where $V$ is a nonempty finite set, $E$ is a finite set and $f$ is a map assigning to every element of $E$ a set of the form $\{u, v\}$ for some $u, v \in V$. The elements of the set $V$ are called vertices, the elements of $E$ edges. Note that we allow edges e for which $f(e)=\{v, v\}$, such edges are called loops. A submultigraph of a multigraph $(V, E, f)$ is a multigraph $\left(V^{\prime}, E^{\prime}, f^{\prime}\right)$ satisfying $V^{\prime} \subseteq V, E^{\prime} \subseteq E, f^{\prime}=\left.f\right|_{E^{\prime}}$.

Definition 1.2.2. Two multigraphs $(V, E, f)$ and $\left(V^{\prime}, E^{\prime}, f^{\prime}\right)$ are isomorphic if there exists a bijection of vertices $\alpha: V \rightarrow V^{\prime}$ and a bijection of edges $\beta: E \rightarrow E^{\prime}$, which for all edges e, $f(e)=\{u, v\}$ satisfy $f^{\prime}(\beta(e))=\{\alpha(u), \alpha(v)\}$. In this thesis, isomorphic graphs will be considered as same.

Definition 1.2.3. A multigraph $(V, E, f)$ is said to be connected if for any two vertices $u, v \in V$ there exists a walk from $u$ to $v$, which means there exists a natural number $n$, vertices $w_{1}, \ldots, w_{n}$ and edges $e_{1}, \ldots, e_{n-1}$ such that $w_{1}=u$, $w_{n}=v$ and $f\left(e_{i}\right)=\left\{w_{i}, w_{i+1}\right\}$.

Definition 1.2.4. An adjacency matrix of a graph $(V, E, f)$ is a square matrix $M$ of order $|V|$, whose rows and columns are indexed by vertices and whose elements are defined as $M_{u, v}=|\{e \in E \mid f(e)=\{u, v\}\}|$.

Remark 1.2.5. A multigraph is (up to isomorphism) uniquely determined by its adjacency matrix.

Definition 1.2.6. Let $M$ be a real matrix of order $n$. Its spectral norm $\|M\|$ is defined as $\max \left\{\left.\frac{\|M v\|}{\|v\|} \right\rvert\, 0 \neq v \in \mathbb{R}^{n}\right\}$, where $\|v\|$ is the Euclidean norm of a vector $v$.

Remark 1.2.7. Since the value $\frac{\|M v\|}{\|v\|}$ depends only on the direction of the vector and not the magnitude, the supremum of the values $\frac{\|M v\|}{\|v\|}$ for nonzero vectors is the same as supremum for only the unit vectors. The set of all unit vectors is compact, therefore the maximum in the definition is attained.

Definition 1.2.8. A simply laced Dynkin diagram is a connected multigraph without loops, whose adjacency matrix has the spectral norm $\|M\|<2$. An extended simply laced Dynkin diagram is a connected multigraph without loops, whose adjacency matrix has the spectral norm $\|M\|=2$. In this thesis, we will call them only Dynkin diagrams, resp. extended Dynkin diagrams.

Proposition 1.2.9. (Perron-Frobenius) Let $M$ be a symmetric matrix with nonnegative elements. Then there exists a nonzero eigenvector $x_{M}$ with nonnegative components corresponding to an eigenvalue $\|M\|$.

Proof. Since $M$ is symmetric, the vector space $\mathbb{R}^{n}$ has an orthogonal basis $B=$ $\left(b_{1}, \ldots, b_{n}\right)$ consisting of eigenvectors of $M, M b_{i}=\lambda_{i} b_{i}$. WLOG suppose $\left|\lambda_{1}\right| \geq$ $\cdots \leq\left|\lambda_{n}\right|$. For a vector $v=\sum a_{i} b_{i}, a_{i} \in \mathbb{R}$ then it holds $\|M v\|^{2}=\left\|\sum a_{i} \lambda_{i} b_{i}\right\|^{2}=$ $\sum a_{i}^{2} \lambda_{i}^{2} \geq \sum a_{i}^{2} \lambda_{n}^{2}=\lambda_{n}^{2}\|v\|$. For the vector $b_{n}$ it holds $\left\|M b_{n}\right\|^{2}=\lambda_{n}^{2}\left\|b_{n}\right\|$, therefore $\left|\lambda_{n}\right|$ is the spectral norm of $M$. Now define a vector $x_{M}$ componentwise as the absolute value of $b_{n}$, i. e. if $b_{n}=\left(\beta_{1}, \ldots, \beta_{n}\right)^{T}$, then $x_{M}=\left|b_{n}\right|=\left(\left|\beta_{1}\right|, \ldots,\left|\beta_{n}\right|\right)^{T}$.

Then compomnentwise $M x_{M}=M\left|b_{n}\right| \geq\left|M b_{n}\right|=\left|\lambda_{n} b_{n}\right|=\|M\| \cdot\left|b_{n}\right|=\| M| | x_{M}$. But this inequality can't be strict in any component, otherwise would $\left\|M x_{M}\right\|$ be greater than $\|M\| \cdot\left\|x_{M}\right\|$, which is a contradiction with the definition of the spectral norm. Therefore $M x_{M}=\|M\| x_{M}$.

Remark 1.2.10. For a general matrix with nonnegative elements, the vector $x_{M}$ does not have to be unique, we will denote by $x_{M}$ any fixed vector with these properties.

Corollary 1.2.11. Let $M$ be an adjacency matrix with nonnegative elements and $x$ its eigenvector corresponding to an eigenvalue $\lambda$, which has only positive components. Then $\|M\|=\lambda$.

Proof. Since $x_{M}$ is nonzero and has only nonnegative components, the value $x^{T} x_{M}$ is positive. Therefore $\lambda x^{T} x_{M}=(M x)^{T} x_{M}=x^{T} M x_{M}=\|M\| x^{T} x_{M}$ implies $\|M\|=\lambda$.

Proposition 1.2.12. Let $G=(V, E, f)$ be a connected multigraph, $M$ its adjacency matrix. Then any nonzero eigenvector $x_{M}$ with nonnegative components has all components strictly positive.

Proof. The vector $x_{M}$ is nonzero and has only nonnegative components, therefore it has some positive component. Suppose there is some vertex, such that the corresponding component of $x_{M}$ is zero. Since $G$ is connected, we can find an edge $e, f(e)=\{u, v\}$, such that the components $x_{M, u}, x_{M, v}$ of vector $x_{M}$ satisfy $x_{M, u}>0$ and $x_{M, v}=0$. Since $M$ and $x_{M}$ are nonnegative, $\left(M x_{M}\right)_{v}=\sum_{x \in V} M_{v, x} x_{M, x} \geq M_{v, u} x_{M, u} \geq x_{M, u}>0$. But $x_{M}$ is an eigenvector of $M$, which leads to a contradiction.

Proposition 1.2.13. Let $G=(V, E, f)$ be a connected multigraph and $M$ its adjacency matrix. Then $x_{M}$ is defined uniquely up to scaling.

Proof. Suppose $x_{M}$ and $x_{M}^{\prime}$ are two eigenvectors with strictly positive components. Since they correspond to the same eigenvalue, any linear combination of them is also an eigenvector. Let $\varepsilon$ be the smallest real number, such that $x_{M}-\epsilon x_{M}^{\prime}$ has a zero component. This vector is an eigenvector with nonnegative components and one of the components is zero, therefore from the previous Proposition it follows it is the zero vector and $x_{M}=\epsilon x_{M}^{\prime}$.

Proposition 1.2.14. Let $G=(V, E, f)$ be a connected multigraph with adjacency matrix $M$ and $G^{\prime}=\left(V^{\prime}, E^{\prime}, f^{\prime}\right)$ its proper connected submultigraph with adjacency matrix $M^{\prime}$. Then $\|M\|>\left\|M^{\prime}\right\|$.

Proof. Define a vector $w \in \mathbb{R}^{V}$ as $w_{u}=x_{M^{\prime}, u}$ if $u \in V^{\prime}$ and $w_{u}=0$ otherwise. Define a matrix $N$ of order $|V|$ as $N_{u, v}=M_{u, v}^{\prime}$ if $u, v \in V^{\prime}$ and 0 otherwise. Since $G$ is connected and $G^{\prime}$ is its proper submultigraph, there must be an edge $e$, $f(e)=\{u, v\}$, such that at least one of the vertices $u, v$ is in $V^{\prime}$, WLOG $u$. Then $(M w)_{v} \geq(N w)_{v}+x_{M^{\prime}, u}>\left\|M^{\prime}\right\| w_{v}$ and for other vertices $i \in V$ it holds $(M w)_{i} \geq$ $(N w)_{i}=\left\|M^{\prime}\right\| w_{i}$, therefore $\|M w\|>\left\|M^{\prime}\right\| \cdot\|w\|$ and $\|M\|>\left\|M^{\prime}\right\|$.

Corollary 1.2.15. An extended Dynkin diagram can't be a proper submultigraph of another extended Dynkin diagram.

Theorem 1.2.16. Every extended Dynkin diagram either is in one of the infinite families $\left\{\tilde{A}_{n} \mid n \geq 1\right\},\left\{\tilde{D}_{n} \mid n \geq 4\right\}$ or is one of three graphs $\tilde{E}_{6}, \tilde{E}_{7}, \tilde{E}_{8}$ :
$\tilde{A}_{1}=\curvearrowleft, \tilde{A}_{2}=\AA, \tilde{A}_{3}=\curvearrowleft, \tilde{A}_{4}=\curvearrowleft, \ldots$

$\tilde{E}_{6}=000 . \tilde{S}_{7}=00, \tilde{E}_{7}=00000000, \tilde{E}_{8}=00000000000$
Proof. For each of the graphs on the list define a vector $x \in \mathbb{R}^{V}$ by assigning these values to vertices:


We can check that for each multigraph, the vector $x$ is an eigenvector of the adjacency matrix $M$ corresponding to the eigenvalue 2 (it can be seen from the picture since for every vertex, the sum of values of its neighbours is twice the value of the vertex, if one counts multiplicities in the multigraph $\tilde{A}_{1}$ ). From Corollary 1.2.11 then follows these multigraphs are extended Dynkin diagrams. Now suppose there exists an extended Dynkin diagram, which is not on the list. Since $\tilde{A}_{1}$ is not its submultigraph, it can't have multiple edges. Since none other $\tilde{A}_{n}$ is its submultigraph, it must be a tree. Since $\tilde{D}_{4}$ is not its submultigraph, its vertices must have a degree at most 3 . Since none other $\tilde{D}_{n}$ is its submultigraph, it must have at most one vertex of degree 3 . On the other hand, it must have a vertex of degree 3 , otherwise, it would be a submultigraph of some $\tilde{A}_{n}$. Therefore it consists of one vertex of degree 3 and three "branches" coming from it:


Since $\tilde{E}_{6}$ is not its submultigraph, the length of one of these "branches" must be 1.


The other two must be longer than 1, otherwise, it would be a submultigraph of some $\tilde{D}_{n}$. On the other hand, one of them must be shorter than 3 , otherwise, $\tilde{E}_{7}$ would be its multigraph.


But then either it is a submultigraph of $\tilde{E}_{8}$ or $\tilde{E}_{8}$ is its submultigraph.
Corollary 1.2.17. Every Dynkin diagram either is in one of the infinite families $\left\{A_{n} \mid n \geq 1\right\},\left\{D_{n} \mid n \geq 4\right\}$ or is one of three graphs $E_{6}, E_{7}, E_{8}$ :
$A_{1}=\circ, A_{2}=0-0, A_{3}=0-0, A_{4}=0-0-0, \ldots$
$D_{4}=0 \oint_{0}^{0}, D_{5}=00 \delta_{0}^{0}, D_{6}=000 \delta_{0}, D_{7}=00000, \ldots$
$E_{6}=000000, E_{7}=0000000, E_{8}=000000000$
Proof. In the proof of the theorem, we actually showed that every connected multigraph either is a submultigraph of an extended Dynkin diagram or contains an extended Dynkin diagram as a submultigraph. Therefore the Dynkin diagrams are exactly connected proper submultigraphs of the extended Dynkin diagrams.

Remark 1.2.18. The extended Dynkin diagrams are called extended because we get them when we "extend" Dynkin diagrams by one new vertex (in a suitable way). The letters $A, D$ and $E$ come from the classification of more general Dynkin diagrams, which uses the letters $B, C, F$ and $G$ for the not-simply-laced ones.

## 2. The algebraic correspondence

### 2.1 Introduction to the representation theory of finite groups

Definition 2.1.1. Let $G$ be a finite group. A representation of $G$ over a field $\mathbb{K}$ is a group homomorphism $\rho: G \rightarrow G L(n, \mathbb{K})$ for some $n \in \mathbb{N}$. The number $n$ is called the dimension of the representation and denoted $\operatorname{dim}(\rho)$. An invariant subspace of the representation is a vector subspace $V \subseteq \mathbb{K}^{n}$, such that for all $g \in G$ and $v \in V$ holds $\rho(g) v \in V$. A representation is said to be irreducible (or simple) if it does not have any invariant subspaces other than 0 and $\mathbb{K}^{n}$. It is said to be faithful if $\rho$ is injective.

Definition 2.1.2. Let $\rho_{1}, \rho_{2}$ be two representations of $G$ over a field $\mathbb{K}$. A homomorphism (resp. isomorphism) of vector spaces $f: \mathbb{K}^{\operatorname{dim}\left(\rho_{1}\right)} \rightarrow \mathbb{K}^{\operatorname{dim}\left(\rho_{2}\right)}$ is said to be a homomorphism (resp. isomorphism) of representations $\rho_{1} \rightarrow \rho_{2}$ if for every $g \in G$ and $v \in \mathbb{K}^{\operatorname{dim}\left(\rho_{1}\right)}$ it holds $f\left(\rho_{1}(g) v\right)=\rho_{2}(g) f(v)$.
Definition 2.1.3. Let $\rho_{1}$, $\rho_{2}$ be two representations of $G$ over a field $\mathbb{K}$. The direct sum $\rho_{1} \oplus \rho_{2}$ is defined as a representation of dimension $\operatorname{dim}\left(\rho_{1}\right)+\operatorname{dim}\left(\rho_{2}\right)$ assigning to $g \in G$ a block diagonal matrix $\left(\begin{array}{cc}\rho_{1}(G) & 0 \\ 0 & \rho_{2}(g)\end{array}\right)$. The tensor product $\rho_{1} \otimes \rho_{2}$ is defined using the tensor product (sometimes called Kronecker product) of matrices as a representation of dimension $\operatorname{dim}\left(\rho_{1}\right) \cdot \operatorname{dim}\left(\rho_{2}\right), g \mapsto \rho_{1}(g) \otimes \rho_{2}(g)$. For iteration of these operations we will write $n \rho=\rho \oplus \cdots \oplus \rho, \rho^{n}=\rho \otimes \cdots \otimes \rho$.

Proposition 2.1.4. Up to isomorphism, the direct sum and the tensor product are both commutative and associative and the tensor product is distributive over the direct sums.

Proof. The proof is done just by rewriting definitions and (in case of commutativities) permuting indices of rows and columns of the resulting matrices.

Theorem 2.1.5. (Mashke) Let $G$ be a finite group and $\mathbb{K}$ be a field, such that char $(\mathbb{K})$ does not divide the order of $G$. Then every representation of $G$ over $\mathbb{K}$ is isomorphic to a direct sum of irreducible representations. This decomposition is unique up to permutation and isomorphism of the irreducible representations.

The proof can be found in Singh 2010 (Chapter 3).
Theorem 2.1.6. Let $G$ be a finite group, $\mathbb{K}$ algebraically closed with characteristic not dividing $|G|$. Then the number of irreducible representations of $G$ over $\mathbb{K}$ (up to isomorphism) is the same as the number of conjugacy classes of $G$.

The proof can be found in Singh 2010 (Chapter 11).
Theorem 2.1.7. Let $G$ be a finite group, $\mathbb{K}$ algebraically closed with characteristic not dividing $|G|$ and $\rho_{1}, \ldots, \rho_{n}$ be representatives of its irreducible representations. Then $\sum_{i=1}^{n} \operatorname{dim}\left(\rho_{i}\right)^{2}=|G|$.

The proof can be found in Singh (2010 (Chapter 11).

Definition 2.1.8. The character $\chi_{\rho}$ of a representation $\rho$ is defined as a map $G \rightarrow \mathbb{K}, g \mapsto \operatorname{tr}(\rho(g))$. The dual representation $\rho^{*}$ is a representation of the same dimension defined as $g \mapsto\left(\rho(g)^{-1}\right)^{T}$

Proposition 2.1.9. Let $\rho_{1}, \rho_{2}$ be two representations of a finite group $G$ over the field $\mathbb{K}$. Then the characters have following properties:

1. $\chi_{\rho_{1} \oplus \rho_{2}}=\chi_{\rho_{1}}+\chi_{\rho_{2}}$
2. $\chi_{\rho_{1} \otimes \rho_{2}}=\chi_{\rho_{1}} \cdot \chi_{\rho_{2}}$
3. $\chi_{\rho_{1}^{*}}(g)=\chi_{\rho_{1}}\left(g^{-1}\right)$
4. If $\mathbb{K}=\mathbb{C}$, $\chi_{\rho_{1}}\left(g^{-1}\right)=\overline{\chi_{\rho_{1}}(g)}$.
5. If $\mathbb{K}=\mathbb{C},\left|\chi_{\rho_{1}}(g)\right| \leq \operatorname{dim}\left(\rho_{1}\right)$.

Proof. The properties 1. and 3. follows straight from the definitions. For property 2. suppose $\rho_{1}(g)$ has elements $d_{1}, \ldots, d_{n}$ on the diagonal and $\rho_{2}(g)$ has elements $e_{1}, \ldots, e_{m}$ on the diagonal. Then the diagonal of $\left(\rho_{1} \otimes \rho_{2}\right)(g)$ consists of the elements of the form $d_{i} e_{j}$, from which it follows that $\chi_{\rho_{1} \otimes \rho_{2}}(g)=\sum_{i, j} d_{i} e_{j}=$ $\left(\sum_{i} d_{i}\right)\left(\sum_{j} e_{j}\right)=\chi_{\rho_{1}}(g) \cdot \chi_{\rho_{2}}(g)$. For the proof of property 4. find a Jordan canonical form of $\rho_{1}(g)$. The elements on the diagonal must be complex numbers $c_{i}$ of the absolute value 1 , otherwise the matrix couldn't satisfy $\rho_{1}(g)^{|G|}=I$. Therefore $\rho_{1}\left(g^{-1}\right)=\rho_{1}(g)^{-1}$ has eigenvalues $c_{i}^{-1}=\overline{c_{i}}$ and $\chi_{\rho_{1}^{*}}(g)=\sum_{i} \overline{c_{i}}=\overline{\chi_{\rho_{1}}(g)}$. 5. follows from triangle inequality and the fact that the eigenvalues of $\rho_{1}(g)$ have absolute value 1 .

Theorem 2.1.10. Let $G$ be a finite group, $\mathbb{K}$ algebraically closed with characteristic not dividing $|G|$ and $\rho_{1}, \ldots, \rho_{n}$ the list of all simple representations up to isomorphism. Then $\sum_{g \in G} \chi_{\rho_{i}}(g) \chi_{\rho_{j}}\left(g^{-1}\right)=\delta_{i j}|G|$.

The proof can be found in Singh (2010 (Chapter 10).
Corollary 2.1.11. Let $\mathbb{K}$ be an algebraically closed field with characteristic zero, $\rho_{1}, \ldots, \rho_{n}$ be the list of irreducible representations up to isomorphism and $\rho$ be any representation with decomposition $\rho=\oplus a_{i} \rho_{i}$. Then $a_{i}=\frac{1}{|G|} \sum_{g \in G} \chi_{\rho_{i}}(g) \chi_{\rho}\left(g^{-1}\right)$.

Proof. $\sum_{g \in G} \chi_{\rho_{i}}(g) \chi_{\rho}\left(g^{-1}\right)=\sum_{j} \sum_{g \in G} \chi_{\rho_{i}}(g) a_{j} \chi_{\rho_{j}}\left(g^{-1}\right)=\sum_{j} \delta_{i j} a_{j}|G|=a_{i}|G|$

Corollary 2.1.12. Let $\mathbb{K}$ algebraically closed with characteristic zero. Then the character uniquely determines the representation (up to isomorphism).

### 2.2 McKay graphs and the algebraic correspondence

In this section, all representations will be considered over $\mathbb{C}$.

Proposition 2.2.1. Let $G$ be a finite group, $\rho_{1}, \ldots, \rho_{n}$ be the list of its irreducible representations up to isomorphism and $\rho$ be a representation satisfying $\rho^{*} \cong \rho$. For $i, j \in\{1, \ldots, n\}$ define numbers $a_{i j}$ as $\rho_{i} \otimes \rho \cong \bigoplus_{j} a_{i j} \rho_{j}$. Then $a_{i j}=a_{j i}$.

Proof. From Corollary 2.1.11 it follows that

$$
\begin{gathered}
a_{i j}=\frac{1}{|G|} \sum_{g \in G} \chi_{\rho_{j}}(g) \chi_{\rho \otimes \rho_{i}}\left(g^{-1}\right)=\frac{1}{|G|} \sum_{g \in G} \chi_{\rho_{j}}(g) \chi_{\rho}\left(g^{-1}\right) \chi_{\rho_{i}}\left(g^{-1}\right)= \\
=\frac{1}{|G|} \sum_{g \in G} \chi_{\rho_{j}}(g) \chi_{\rho}(g) \chi_{\rho_{i}}\left(g^{-1}\right)=\frac{1}{|G|} \sum_{h \in G} \chi_{\rho \otimes \rho_{j}}\left(h^{-1}\right) \chi_{\rho_{i}}(h)=a_{j i},
\end{gathered}
$$

where the middle equality on the second line is using a substitution $h:=g^{-1}$.
Definition 2.2.2. Let $G$ be a finite group. A self-dual representation is a representation satisfying $\rho \cong \rho^{*}$. The McKay graph of $G$ and self-dual $\rho$ is a multigraph, whose vertices are the irreducible representation of $G$ and its adjacency matrix consists of the coefficients $a_{i j}$.

Proposition 2.2.3. Let $(V, E, f)$ be a McKay graph of a group $G$ and a self-dual representation $\rho$, let $M$ be its adjacency matrix. Then $\|M\|=\operatorname{dim}(\rho)$.

Proof. Define a vector $x \in \mathbb{R}^{V}$ componentwise as $x_{i}=\operatorname{dim}\left(\rho_{i}\right)$. For any $i \in$ $\{1, \ldots, n\}$ from $\rho_{i} \otimes \rho \cong \oplus_{j} a_{i j} \rho_{j}$ it follows that $\operatorname{dim}\left(\rho_{i}\right) \cdot \operatorname{dim}(\rho)=\sum_{j} a_{i j} \operatorname{dim}\left(\rho_{j}\right)$, therefore $x_{i} \operatorname{dim}(\rho)=\sum_{j} a_{i j} \operatorname{dim}\left(\rho_{j}\right)$ and therefore $\operatorname{dim}(\rho) x=M x$. The statement then follows from Corollary 1.2.11.

Definition 2.2.4. Let $G$ be a finite group. Its trivial representation $\rho_{\text {triv }}$ is the constant homomorphism $G \rightarrow G L(1, \mathbb{C}), g \mapsto 1$. Let $G$ be a finite subgroup of $S L(2, \mathbb{C})$. Its canonical representation is the embedding $G \rightarrow G L(2, \mathbb{C}), g \mapsto g$.

Theorem 2.2.5. (Burnside) Let $G$ be a finite group, $\rho$ a faithful representation of $G$ and $\rho_{i}$ an irreducible representation. Then there exists $n_{i} \in \mathbb{N}$ such that $\rho_{i}$ is contained in the decomposition of $\rho^{n_{i}}$ into irreducible representations.
Proof. In other words, we want to show that $\sum_{g \in G} \chi_{\rho_{i}}(g)\left(\overline{\chi_{\rho}(g)}\right)^{n_{i}}>0$ for some $n_{i} \in \mathbb{N}$. Lets assume the converse, $\forall n \in \mathbb{N}: \sum_{g \in G} \chi_{\rho_{i}}(g)\left(\overline{\chi_{\rho}(g)}\right)^{n}=0$. Since all group elements are of finite order, the matrices $\rho(g)$ are diagonalizable and have eigenvalues of the form $e^{\frac{2 k \pi i}{|G|}}$, therefore $\left|\chi_{\rho}(g)\right| \leq \operatorname{dim}(\rho)$. Let $\gamma=\{g \in$ $G\left|\operatorname{dim}(\rho)=\left|\chi_{\rho}(g)\right|\right\}$. Then for $g \in \gamma, \rho(g)$ must have only one eigenvalue, otherwise the triangle inequality $\left|\lambda_{1}+\cdots+\lambda_{\operatorname{dim}(\rho)}\right| \leq 1+\cdots+1=\operatorname{dim}(\rho)$ would be strict. Therefore $\rho(g)=\lambda_{g} I$ for some $\lambda_{g}=e^{\frac{2 k \pi i}{\mid G T}}$. From $\rho$ being faithful it follows that for distinct elements of $\gamma$ are the numbers $\lambda_{g}$ distinct. Let us now choose $k \in \mathbb{Z}$ and define a sequence of functions $f_{k, n}: G \rightarrow \mathbb{C}, g \mapsto \frac{\left.\chi_{\rho}(g)\right|^{|G| n+k}}{\operatorname{dim}(\rho)^{|G| n+k}}$. As $n \rightarrow \infty$, this sequence is constant at $g \in \gamma$ and decreases exponentially (in the absolute value) at $g \notin \gamma$, therefore it has a pointwise limit $f_{k}$, which has the value $\lambda_{g}^{k}$ at $g \in \gamma$ and 0 otherwise. When we view maps $G \rightarrow \mathbb{C}$ as vectors in $\mathbb{C}^{G}$, the assumption $\sum_{g \in G} \chi_{\rho_{i}}(g)\left(\overline{\chi_{\rho}(g)}\right)^{n}=0$ tells us that $f_{k, n}$ and $\chi_{\rho_{i}}$ are orthogonal. From continuity of the dot product it follows that $f_{k}$ and $\chi_{\rho_{i}}$ are orthogonal. For a vector $v \in \mathbb{C}^{G}$ define its restriction $\tilde{v} \in \mathbb{C}^{\gamma}$ by omitting the components
outside $\gamma$. Since $f_{k}$ are zero outside $\gamma$, it means that $\tilde{f}_{k}$ and $\tilde{\chi}_{\rho_{i}}$ are orthogonal. If we name the elements of $\gamma$ as $g_{1}, \ldots, g_{|\gamma|}$, we can combine the equations for $k \in\{0, \ldots,|\gamma|-1\}$ into one matrix equation.

$$
\left(\begin{array}{ccccc}
1 & 1 & 1 & \ldots & 1 \\
\lambda_{g_{1}} & \lambda_{g_{2}} & \lambda_{g_{3}} & \ldots & \lambda_{g_{\gamma \mid}} \\
\lambda_{g_{1}}^{2} & \lambda_{g_{2}}^{2} & \lambda_{g_{3}}^{2} & \ldots & \lambda_{g_{|\gamma|}}^{2} \\
\vdots & \vdots & \vdots & \ddots & \vdots \\
\lambda_{g_{1}}^{|\gamma|-1} & \lambda_{g_{2}}^{|\gamma|-1} & \lambda_{g_{3}}^{|\gamma|-1} & \ldots & \lambda_{g_{|\gamma|}}^{|\gamma|-1}
\end{array}\right) \cdot \overline{\left(\begin{array}{c}
\chi_{\rho_{i}}\left(g_{1}\right) \\
\chi_{\rho_{i}}\left(g_{2}\right) \\
\chi_{\rho_{i}}\left(g_{3}\right) \\
\vdots \\
\chi_{\rho_{i}}\left(g_{|\gamma|}\right)
\end{array}\right)}=\left(\begin{array}{c}
0 \\
0 \\
0 \\
\vdots \\
0
\end{array}\right)
$$

In the matrix on the left we recognise the Vandermonde matrix, which is regular, since $\lambda_{g_{i}}$ are pairwise. This means that $\chi_{\rho_{i}}$ is zero on $\gamma$, but since the neutral element $e$ is in $\gamma$ and $\chi_{\rho_{i}}(e)=\operatorname{dim}\left(\rho_{i}\right)$, this is a contradiction.

Corollary 2.2.6. Let $G$ be a finite group and $\rho$ a self-dual faithful representation. Then the McKay graph of $G$ and $\rho$ is connected.

Proof. By induction, we will prove that if the decomposition of $\rho^{n}$ contains irreducible $\tilde{\rho}$, then there exist a walk from $\rho_{\text {triv }}$ to $\tilde{\rho}$. For $n=1$ the statement holds, since if $\tilde{\rho}$ is contained in $\rho=\rho \otimes \rho_{\text {triv }}$, it is a neighbor of $\rho_{\text {triv }}$ by definition of the McKay graph. Suppose the statement holds for $n$ and $\tilde{\rho}$ is contained in $\rho^{n+1}$. When we decompose $\rho^{n}$ as $\bigoplus a_{i} \rho_{i}$, we learn that $\tilde{\rho}$ is contained in $\oplus a_{i} \rho_{i} \otimes \rho$. From uniqueness of the decomposition it must be contained in some $\rho_{i} \otimes \rho$ such that $a_{i}>0$. When we take a walk from $\rho_{\text {triv }}$ to $\rho_{i}$ and extend it by a vertex $\tilde{\rho}$ (and some suitable edge between $\rho_{i}$ and $\tilde{\rho}$ ), we get a walk to $\tilde{\rho}$. Now, to get a walk between general vertices $u$, $v$, one can take a walk from $u$ to $\rho_{\text {triv }}$ and follow it by a walk from $\rho_{\text {triv }}$ to $v$.

Proposition 2.2.7. Let $G$ be a finite group, $\rho$ be an irreducible representation of $G$ and $g$ be an element of the center of $G$. Then $\rho(g)=\lambda I$ for some $\lambda \in \mathbb{C}$.

Proof. We know that $\rho(g)$ is diagonalizable, let $\lambda$ be an eigenvalue and $V$ the corresponding eigenspace. We want to show that $V$ is an invariant subspace of $\rho$. Let $v \in V, h \in G$. Then $\rho(g) \rho(h) v=\rho(h) \rho(g) v=\lambda \rho(h) v$, therefore $\rho(h) v \in V$. Since $\rho$ is irreducible, this means $V=\mathbb{C}^{\operatorname{dim}(\rho)}$.

Corollary 2.2.8. Let $G$ be a finite abelian group. Then all of its irreducible representations are one-dimensional.

Proof. Since all matrices $\rho(g)$ are of the form $\lambda I$, any subspace of $\mathbb{C}^{\operatorname{dim}(\rho)}$ is invariant.

Proposition 2.2.9. Let $G$ be a finite group with a nontrivial center, $\rho_{i}$ an irreducible representation and $\rho_{j}$ a faithful irreducible representation. Then $\rho_{i}$ does not appear in the decomposition of $\rho_{j} \otimes \rho_{i}$.

Proof. Suppose it does. Let $g \neq e$ be an element of the center. From Proposition 2.2.7 it follows that for any irreducible $\rho_{k}, \chi_{\rho_{k}}(g)=\lambda_{k} \operatorname{dim}\left(\rho_{k}\right)$ for some complex unit $\lambda_{k}$. Let $\oplus a_{k} \rho_{k}$ be the decomposition of $\rho_{j} \otimes \rho_{i}$. When we apply the absolute value to the equality

$$
\begin{gathered}
\lambda_{i} \lambda_{j} \sum a_{k} \operatorname{dim}\left(\rho_{k}\right)=\lambda_{i} \lambda_{j} \sum a_{k} \chi_{\rho_{k}}(e)=\lambda_{i} \lambda_{j} \chi_{\rho_{j} \otimes \rho_{i}}(e)= \\
=\lambda_{i} \lambda_{j} \operatorname{dim}\left(\rho_{j}\right) \cdot \operatorname{dim}\left(\rho_{i}\right)=\chi_{\rho_{j} \otimes \rho_{i}}(g)=\sum a_{k} \chi_{\rho_{k}}(g)=\sum a_{k} \lambda_{k} \operatorname{dim}\left(\rho_{k}\right),
\end{gathered}
$$

we get $\sum a_{k} \operatorname{dim}\left(\rho_{k}\right)=\left|\sum a_{k} \lambda_{k} \operatorname{dim}\left(\rho_{k}\right)\right|$. On the other hand, the statement $\sum a_{k}\left|\lambda_{k}\right| \operatorname{dim}\left(\rho_{k}\right) \leq\left|\sum a_{k} \lambda_{k} \operatorname{dim}\left(\rho_{k}\right)\right|$ follows from the triangle equality, which is an equality if and only if all of the nonzero summands have the same argument $\arg (z)=\frac{z}{|z|}$. This means $\lambda_{k}$ is the same for all $k$ satisfying $a_{k}>0$. Since there exists a loop from $\rho_{i}$ to $\rho_{i}$, we know that $a_{i}>0$, therefore

$$
\lambda_{i} \lambda_{j} \sum a_{k} \operatorname{dim}\left(\rho_{k}\right)=\sum a_{k} \lambda_{k} \operatorname{dim}\left(\rho_{k}\right) .
$$

We can cancel out $\lambda_{i} \sum a_{k} \operatorname{dim}\left(\rho_{k}\right)$ in and get $\lambda_{j}=1$, which is in contradiction with $\rho_{j}$ being faithful.

Proposition 2.2.10. Let $G$ be a nontrivial finite subgroup of $S L(2, \mathbb{C})$ and $\rho$ be its canonical representation. Then $\rho$ is self-dual and the McKay graph of $G$ and $\rho$ has no loops.

Proof. Since for every $g \in G$ the eigenvalues of $\rho(g)$ are complex conjugates of each other, the character $\chi_{\rho}$ is real-valued. Therefore $\chi_{\rho^{*}}=\chi_{\rho}$ and from Corollary 2.1.12 $\rho \cong \rho^{*}$. If $\rho$ is irreducible, then the proof follows from Proposition 2.2.9 (all nontrivial finite subgroups of $S L(2, \mathbb{C})$ have a nontrivial center). If not, it decomposes into two one-dimensional representations $\rho_{1}, \rho_{2}$. Since eigenvalues of matrices in $S L(2, \mathbb{C})$ are inverse to each other, $\rho_{1}$ and $\rho_{2}$ must be dual. From this follows that both $\rho_{1}$ and $\rho_{2}$ are faithful, since if $g \in \operatorname{Ker}\left(\rho_{1}\right)$, then $g \in \operatorname{Ker}\left(\rho_{2}\right)$ and $g \in \operatorname{Ker}\left(\rho_{1} \oplus \rho_{2}\right)$, which contradicts the fact that canonical representation is by definition faithful. If $\rho_{i}$ is now any irreducible representation, it does not appear in the decomposition of $\left(\rho_{1} \otimes \rho_{i}\right) \oplus\left(\rho_{2} \otimes \rho_{i}\right) \cong \rho \otimes \rho_{i}$, which is what we wanted to prove.

Corollary 2.2.11. Let $G$ be a nontrivial finite subgroup of $S L(2, \mathbb{C})$ and $\rho$ its canonical representation. Then the McKay graph of $G$ and $\rho$ is an extended Dynkin diagram.

Proof. Follows from Proposition 2.2.3. Corollary 2.2.6 and Proposition 2.2.10
Proposition 2.2.12. Let $G$ be a nontrivial finite subgroup of $S L(2, \mathbb{C})$. Then the dimension vector $x$ used in the proof of Proposition 2.2.3 is the same as the eigenvector $x_{M}$ for the corresponding extended Dynkin diagram used in the proof of Theorem 1.2.16:




Proof. Both of these vectors are eigenvectors of the adjacency matrix with positive components, from Proposition 1.2.13 follows $x=\alpha x_{M}$ for some $\alpha \in \mathbb{R}$. Since components of $x$ are positive integers, $\alpha \in \mathbb{N}$. Since the trivial representation is an irreducible representation with dimension $1, \alpha \leq 1$.

Theorem 2.2.13. The extended Dynkin diagrams corresponding to nontrivial finite subgroups of $S L(2, \mathbb{C})$ are:

| Group: | Diagram: |
| :---: | :---: |
| $C_{n}$ | $\tilde{A}_{n-1}$ |
| $B D_{4 n}$ | $\tilde{D}_{n+2}$ |
| $B T$ | $\tilde{E}_{6}$ |
| $B O$ | $\tilde{E}_{7}$ |
| $B I$ | $\tilde{E}_{8}$ |

Proof. From Theorem 2.1.6 the diagram determines the number of conjugacy classes, from Proposition 2.2.12 and Theorem 2.1.7 it determines the order of the group:

| Diagram: | $\|G\|$ | Conjugacy classes |
| :---: | :---: | :---: |
| $\tilde{A}_{n-1}$ | $n$ | $n$ |
| $\tilde{D}_{n+2}$ | $4 n$ | $n+3$ |
| $\tilde{E}_{6}$ | 24 | 7 |
| $\tilde{E}_{7}$ | 48 | 8 |
| $\tilde{E}_{8}$ | 120 | 9 |

Therefore for a given nontrivial finite subgroup of $S L(2, \mathbb{C})$, the diagram in the statement of the Theorem is the only possible one.

## 3. The geometric correspondence

### 3.1 Introduction to algebraic geometry

In this chapter, the reader is expected to know the basics of commutative algebra and algebraic geometry. All algebras in this chapter are considered associative, commutative and unital and over the field $\mathbb{C}$.

Definition 3.1.1. By the $n$-dimensional affine space $\mathbb{A}^{n}$ we mean $\mathbb{C}^{n}$ as a set (later equipped with a topology). Let $S \subseteq \mathbb{C}\left[x_{1}, \ldots, x_{n}\right]$. The common zero set of $S$ is the set $V(S):=\left\{a \in \mathbb{A}^{n} \mid \forall f \in S: f(a)=0\right\}$. Subsets of $\mathbb{A}^{n}$ of the form $V(S)$ are called affine algebraic sets. Let $X \subseteq \mathbb{C}\left[x_{1}, \ldots, x_{n}\right]$. Then the ideal of $X$ is the ideal $I(X):=\left\{f \in \mathbb{A}\left[x_{1}, \ldots, x_{n}\right] \mid \forall x \in X: f(x)=0\right\}$.

Proposition 3.1.2. The intersections and finite unions of algebraic sets are algebraic sets. The empty set and the set $\mathbb{A}^{n}$ are algebraic sets. The complements of algebraic sets therefore form a topology on $\mathbb{A}^{n}$.

Proof. Straight from the definition follows $V\left(\cup S_{i}\right)=\bigcap V\left(S_{i}\right)$ and $V\left(S_{1} S_{2}\right)=$ $V\left(S_{1}\right) \cup V\left(S_{2}\right)$ (where $S_{1} S_{2}=\left\{s t \mid s \in S_{1}, t \in S_{2}\right\}$ ). The empty set is equal to $V(1)$ and the set $\mathbb{C}^{n}$ is equal to $V(0)$.

Definition 3.1.3. The topology from the previous Proposition and the induced topologies on subsets of $\mathbb{C}^{n}$ are called the Zariski topology. Unless stated otherwise, open and closed sets will be from now on considered with respect to the Zariski topology.

Definition 3.1.4. Let $(X, \tau)$ be a topological space. If for any two closed subsets $X_{1}, X_{2}$ satisfying $X_{1} \cup X_{2}=X$ holds $X_{1}=X$ or $X_{2}=X, X$ is said to be irreducible (alternatively, if every nonempty open of $X$ subset is dense in $X$ ). A subset of $X$ is said to be irreducible if it is irreducible as a topological space with the induced topology. Nonempty irreducible affine algebraic sets are called affine varieties.

Definition 3.1.5. Let $X \subseteq \mathbb{A}^{n}$ be an affine algebraic set. The coordinate algebra of $X$ is the algebra $\mathbb{C}[X]=\mathbb{C}\left[x_{1}, \ldots, x_{n}\right] / I(X)$ (which can be considered as an algebra of polynomial functions $X \rightarrow \mathbb{C}$ ).

Definition 3.1.6. Let $X \subseteq \mathbb{A}^{m}, Y \subseteq \mathbb{A}^{n}$ be affine algebraic sets. $A$ map $X \rightarrow Y$, $\left(x_{1}, \ldots, x_{m}\right) \mapsto\left(f_{1}\left(x_{1}, \ldots, x_{m}\right), \ldots, f_{n}\left(x_{1}, \ldots, x_{m}\right)\right)$ is a morphism of varieties if the functions $f_{1}, \ldots, f_{n}$ are polynomial.

Example. As a set, the coordinate algebra of $X$ is therefore the set of all morphisms $X \rightarrow \mathbb{A}^{n}$.

Theorem 3.1.7. The coordinate algebra of an affine algebraic set is finitely generated and does not contain any nontrivial nilpotent elements. This assignment between the category of affine algebraic sets and the category of finitely algebras without nontrivial nilpotents is a contravariant equivalence of categories, which maps the morphism of varieties $f: X \rightarrow Y$ to a morphism of the coordinate algebras $\tilde{f}: \mathbb{C}[Y] \rightarrow \mathbb{C}[X], g \mapsto g \circ f$.

Proof. The coordinate algebra is finitely generated, because it is a quotient algebra of finitely generated $\mathbb{C}\left[x_{1}, \ldots, x_{n}\right]$. It does not contain nontrivial nilpotents, since it consists of valued functions with operations defined pointwise. The proof of equivalence of categories can be found in Wallach 2017 (section 1.2.1) or Smith et al. 2000 (section 2.5). Both of these books differ in the terminology, their definition of an affine variety correspond to our definition of an affine algebraic set. In the book Wallach [2017], the symbol $\mathcal{O}(Z)$ stands for the local ring of a set $Z \subseteq \mathbb{A}^{n}$, which for closed sets is the same as the coordinate algebra.

### 3.2 Geometric invariant theory

Definition 3.2.1. Let $G$ be a finite group and $X$ be an affine variety. By a group action on $X$ we mean a group homomorphism $\varphi: G \rightarrow A u t(X)$. The equivalence of the categories lets us view this as a group homomorphism $\tilde{\varphi}: G \rightarrow \operatorname{Aut}(\mathbb{C}[X])$, $\tilde{\varphi}_{g}(f)=f \circ \varphi_{g^{-1}}$. The algebra of invariant polynomials is the algebra $\mathbb{C}[X]^{G}=$ $\left\{f \in \mathbb{C}[X] \mid \forall g \in G: \tilde{\varphi}_{g}(f)=f\right\}$. If this algebra is finitely generated, we denote the corresponding affine variety $X / / G$ and call it the GIT-quotient of $X$. The equivalence of categories then lets us view the inclusion $i: \mathbb{C}[X]^{G} \rightarrow \mathbb{C}[X]$ as a morphism $\pi: X \rightarrow X / / G$.

Theorem 3.2.2. Let $G \subseteq G L(n, \mathbb{C})$ be a finite group acting on $\mathbb{A}^{n}$ by multiplying the coordinate vector:

$$
\varphi_{g}\left(\begin{array}{c}
x_{1} \\
\vdots \\
x_{n}
\end{array}\right)=g \cdot\left(\begin{array}{c}
x_{1} \\
\vdots \\
x_{n}
\end{array}\right)
$$

Then $\mathbb{C}[X]^{G}$ is finitely generated. Moreover, the ideal $I \subseteq \mathbb{C}[X]$ generated by all homogeneous invariant polynomials is generated by finitely many homogeneous invariant polynomials and these polynomials also generate $\mathbb{C}[X]^{G}$.

Proof. The proof can be found in Wallach 2017 (section 3.1.4).
Proposition 3.2.3. Let $G$ be the group $C_{n}<S L(2, \mathbb{C})$. Then the affine variety $\mathbb{A}^{2} / / G$ is isomorphic to $V\left(a b-c^{n}\right) \in \mathbb{A}^{3}$.

Proof. WLOG assume $C_{n}=\langle M\rangle$, where $M=\left(\begin{array}{cc}e^{\frac{2 \pi i}{n}} & 0 \\ 0 & e^{-\frac{2 \pi i}{n}}\end{array}\right)$. A polynomial is invariant, if and only if it is invariant under the action of the generators of the group. Since $M$ is diagonal, its action on the monomials changes only the coeficient of the monomial. A polynomial is therefore invariant if and only if all of its monomials are invariant. The matrix $M$ sends $x^{k} y^{l}$ to $\left(e^{-\frac{2 \pi i}{n}} x\right)^{k}\left(e^{\frac{2 \pi i}{n}}\right)^{l}=x^{k} y^{l} e^{\frac{2 \pi i(l-k)}{n}}$, therefore the monomial is invariant if and only if $k \equiv l(\bmod n)$. The ideal $I$ from the previous Theorem is therefore $\left(x^{n}, y^{n}, x y\right)$ (since these polynomials are invariant and any nonconstant invariant polynomial is divisible by at least one of them). Let $\psi: \mathbb{C}[a, b, c] \rightarrow \mathbb{C}[X]^{G}$ be a homomorphism defined by $\psi(a)=x^{n}$, $\psi(b)=y^{n}, \psi(c)=x y$. Since the generators of $\mathbb{C}[X]^{G}$ are in the image, the homomorphism is surjective, therefore $\mathbb{C}[X]^{G} \cong \frac{\mathbb{C}[a, b, c]}{\operatorname{Ker}(\psi)}$. Clearly $a b-c^{n} \in \operatorname{Ker}(\psi)$, therefore $\mathbb{A} / / G$ is some subvariety of $V\left(a b-c^{n}\right)$. The projection morphism is defined as $(x, y) \mapsto\left(x^{n}, y^{n}, x y\right)$. If for $(a, b, c) \in V\left(a b-c^{n}\right)$ holds $a \neq 0$, then
$\pi\left(\sqrt[n]{a}, \frac{c}{\sqrt[n]{a}}\right)=(a, b, c)$. If $a=0$, then $\pi(0, \sqrt[n]{b})=(a, b, c)$. Therefore $\pi$ is surjective, therefore $\operatorname{Ker}(\psi)=V\left(a b-c^{n}\right)$ and $\mathbb{A} / / G$ is the whole affine variety $V\left(a b-c^{n}\right)$.

### 3.3 Singularities

Definition 3.3.1. Let
Some difficulties arise when one is working with singular varieties. To avoid these problems, we may want to find some different variety, which is nonsingular, but still carries some information about the original variety. It can also help us better understand the singularities. The simplest method of finding these resolutions is blowing up points. However, the result of it usually isn't an affine variety (even if the original variety is), but a more general object. This object can be embedded into the projective space, however, we will follow a more abstract approach and think of it as a topological space (with some additional structure), which is covered by affine varieties, analogously to the way how differential varieties are covered by Euclidianly open subsets of $\mathbb{R}^{n}$. We will call these objects abstract varieties or just varieties. The precise definition can be found in Wallach 2017 (section 1.4.3)

Definition 3.3.2. Let $X$ be a variety. A resolution of singularities is a nonsingular variety $Y$ together with a morphism $f: Y \rightarrow X$, such that:

- The morphism $f$ is proper, i.e. if we equip $X$ and $Y$ with Euclidean topology, the preimage of a compact set is compact.
- When restricted, $f$ is an isomorphism of varieties $f^{-1}\left(X^{\text {reg }}\right) \rightarrow X^{\text {reg }}$, where $X^{\text {reg }}$ is the open subvariety of $X$ consisting of all regular points.

An exceptional locus of the resolution is the set $f^{-1}\left(X^{\text {sing }}\right)$, where $X^{\text {sing }}$ is the set of all singular points $X \backslash X^{\text {reg }}$.

Remark 3.3.3. If $X \subseteq \mathbb{A}^{n}$ is an affine variety, it can be equipped with the Euclidean topology induced from the by the inclusion $X \rightarrow \mathbb{C}^{n}$. For abstract varieties, the topology can be constructed locally the same way. More about this in Neeman (2007) (chapter 4).

Definition 3.3.4. Let $p$ be a point of an affine variety $X \subseteq \mathbb{A}^{n}$. The blowup of $X$ at the point $p$ is the closure of

$$
\Gamma=\left\{(x, y) \in(X \backslash\{p\}) \times \mathbb{P}^{n-1} \mid\left[x_{1}-p_{1}: \ldots: x_{n}-p_{n}\right]=\left[y_{1}: \ldots: y_{n}\right]\right\}
$$

in $X \times \mathbb{P}^{n-1}$. This variety is denoted $B l_{p}(X)$ and there is a canonical morphism $\pi: B l_{p}(X) \rightarrow X,(x, y) \mapsto x$. Again, the set $\pi^{-1}(\{p\})$ is called the exceptional locus.

Remark 3.3.5. A blowup can be defined even for abstract varieties as in Hauser [2014]. For our purposes suffices to know that the blowup of an abstract variety, looks on the affine cover as the blowup of affine variety if the covering affine variety contains the point $p$ and is an isomorphism otherwise.

Example. Let $X=V\left(x_{1}^{3}+x_{1}^{2}-x_{2}^{2}\right) \subseteq \mathbb{A}^{2}$ and $p=(0,0)$. We can cover the projective space by two open subsets, on which $y_{1} \neq 0$, resp. $y_{2} \neq 0$. Since the closure commutes with finite unions, we can use the closures of the sets in the open cover to find the closure of $\Gamma$. The first can be parameterized by $x_{1}, x_{2}$ and $\tilde{x}_{2}$, where $\left[y_{1}: y_{2}\right]=\left[1: \tilde{x}_{2}\right]$. The corresponding subset of $\Gamma$ is therefore parameterized by

$$
\begin{aligned}
& \left\{\left(x_{1}, x_{2}, \tilde{x}_{2}\right) \in \mathbb{A}^{3} \mid x_{1} \neq 0, x_{1}^{3}+x_{1}^{2}-x_{2}^{2}=0, x_{2}=x_{1} \tilde{x}_{2}\right\}= \\
& =\left\{\left(x_{1}, x_{2}, \tilde{x}_{2}\right) \in \mathbb{A}^{3} \mid x_{1} \neq 0, x_{1}+1-\tilde{x}_{2}^{2}=0, x_{2}=x_{1} \tilde{x}_{2}\right\} .
\end{aligned}
$$

By taking the closure of this set in $\left\{(x, y) \in X \times \mathbb{P}^{n-1} \mid y_{1} \neq 0\right\}$, we remove the condition $x_{1} \neq 0$ and therefore get two new points points $\left(x_{1}, x_{2}, \tilde{x}_{2}\right)=(0,0, \pm 1)$. Since $x_{2}$ can be expressed from $x_{1}$ and $\tilde{x}_{2}$, it is not needed for the parameterization and the covering affine variety is isomorphic to $V\left(x_{1}+1-\tilde{x}_{2}^{2}\right) \subseteq \mathbb{A}^{2}$.
We could similarly parameterize the open set $y_{2} \neq 0$, however, in this particular example it is not necessary, since the first parameterization covers the whole $\Gamma$. The morphism $\pi: B l_{p}(X) \rightarrow X$ is given by $\left(x_{1}, \tilde{x}_{2}\right) \mapsto\left(x_{1}, x_{2}\right)=\left(x_{1}, x_{1} \tilde{x}_{2}\right)$

This process can be generalized to an algorithm for finding affine cover of any blowup of hyperplane, i.e. any blowup $B l_{p}(X)$ of a variety of the form $X=$ $V(f) \subseteq \mathbb{A}^{n}$.

- cover $\mathbb{P}^{n-1}$ by $n$ open subsets $U_{i}$ defined by $y_{i} \neq 0$.
- parametrize every $X \times U_{i}$ by $x_{1}, \ldots, x_{n}, x_{1, i}, \ldots, x_{i-1, i}, x_{i+1, i}, \ldots, x_{n, i}$, where $\left[y_{1}: \cdots: y_{n}\right]=\left[x_{1}-p_{1}: \cdots: x_{i-1}-p_{i-1}: 1: x_{i+1}-p_{i+1}: \ldots, x_{n}-p_{n}\right]$.
- express the coordinates $x_{1}, \ldots, x_{i-1}, x_{i+1}, \ldots, x_{n}$ in $f$ using the coordinates $x_{1, i}, \ldots, x_{i-1, i}, x_{i}, x_{i+1, i} \ldots, \ldots, x_{n, i}$.
- divide the transformed $f \in \mathbb{C}\left[x_{1, i}, \ldots, x_{i-1, i}, x_{i}, x_{i+1, i} \ldots, \ldots, x_{n, i}\right]$ by the highest possible power of $x_{i}-p_{i}$
- find the closure in $X \times U_{i}$, i.e. add the points with $x_{i}=p_{i}$
- on the intersections of multiple maps, find the transition map

Proposition 3.3.6. The restriction of the blowup morphism is an isomorphism of the varieties $\pi^{-1}(X \backslash\{p\}) \rightarrow X \backslash\{p\}$
Proof. The set $\Gamma \cup\left(p \times \mathbb{P}^{n-1}\right)$ is closed, therefore this restriction of $\pi$ is injective. The map
$X \backslash\{p\} \rightarrow(X \backslash\{p\}) \times \mathbb{P}^{n-1},\left(x_{1}, \ldots, x_{n}\right) \mapsto\left(x_{1} \ldots, x_{n},\left[x_{1}-p_{1}: \ldots: x_{n}-p_{n}\right]\right)$
is the inverse morphism.
Proposition 3.3.7. The blowup morphism is proper.
Proof. Let $C \subseteq X$ be a compact subset of $X$ (with respect to Euclidean topology) Since $X$ is covered by affine varieties, we can find a compact neighborhood $U$ of $p \in X$. Let $V=X \backslash U$. Since on $(X \backslash\{p\}) \times \mathbb{P}^{n-1}$ is $\pi$ an isomorphism, $V \cap C$ is homeomorphic to $\pi^{-1}(V \cap C)$, therefore $\pi^{-1}(C \cap V)$ is compact. Since $\mathbb{P}^{n-1}$ is compact (Wallach 2017, Lemma 1.29), the set $U \times \mathbb{P}^{n-1}$, therefore the closed subset $\pi^{-1}(U \cap C)$ is also compact. Therefore $\pi^{-1}(C)=\pi^{-1}(U \cap C) \cup \pi^{-1}(V \cap C)$ is compact.

Corollary 3.3.8. We can construct a sequence $X_{n} \rightarrow X_{n-1} \rightarrow \cdots \rightarrow X_{1} \rightarrow$ $X_{0}=X$, in which $X_{k+1}$ is a blowup of $X_{k}$ at some singular point. If $X_{n}$ is a nonsingular variety, then the composition of the blowup morphisms $X_{n} \rightarrow X$ is a resolution of singularities.

Theorem 3.3.9. Let $G$ be a nontrivial finite subgroup of $S L(2, \mathbb{C})$ and $X=$ $\mathbb{A}^{2} / / G$. Then we can get a resolution of $X$ by consequent blowups at singular points. If we then construct a multigraph, whose vertices are the irreducible components of the exceptional locus and the number of edges between components $U$ and $V$ is $|U \cap V|$, we get a non-extended version of the extended Dynkin diagram from Theorem 2.2.13.

Proof. We will show this for each group separately.
$\mathbf{C}_{2}$ : The variety $X=\mathbb{A}^{2} / / G$ is isomorphic to $V\left(x y-z^{2}\right)$. The only singular point is the point $p=(0,0,0)$. The blowup is covered by affine varieties $\overline{U_{i}}$. Using the algorithm at the beginning of the section, we get the affine cover of $B l_{p}(X)$ :

- $B l_{p}(X) \cap\left(X \times U_{x}\right)$ is covered by $V\left(\tilde{y}-\tilde{z}^{2}\right)$ with the morphism $\pi_{x}:(x, \tilde{y}, \tilde{z}) \mapsto$ $(x, \tilde{y} x, \tilde{z} x,[1: \tilde{y}: \tilde{z}])$.
- $B l_{p}(X) \cap\left(X \times U_{y}\right)$ is covered by $V\left(\tilde{x}-\tilde{z}^{2}\right)$ with the morphism $\pi_{y}:(\tilde{x}, y, \tilde{z}) \mapsto$ $(\tilde{x} y, y, \tilde{z} y,[\tilde{x}: 1: \tilde{z}])$.
- $B l_{p}(X) \cap\left(X \times U_{y}\right)$ is covered by $V(\tilde{x} \tilde{y}-1)$ with the morphism $\pi_{z}:(\tilde{x}, \tilde{y}, z) \mapsto$ $(\tilde{x} z, \tilde{y} z, z,[\tilde{x}: \tilde{y}: 1])$.

All three varieties $V\left(\tilde{y}-\tilde{z}^{2}\right), V\left(\tilde{x}-\tilde{z}^{2}\right)$ and $V(\tilde{x} \tilde{y}-1)$ are nonsingular, therefore $B l_{p}(X)$ is nonsingular, therefore it is a resolution of singularities. We will now study the exceptional locus, i.e. $\pi^{-1}(p)$. In the first set of the affine cover, the preimage of $p=(0,0,0)$ consist of the points parameterized by $(x, \tilde{y}, \tilde{z})=$ $\left(0, \tilde{z}^{2}, \tilde{z}\right)$, which is an irreducible set. Similarly for $U_{y}, \pi^{-1}(0,0,0)$ is parameterized by $y=0$ and $\tilde{x}=\tilde{z}^{2}$ and for $U_{z}$ it is parameterized by $z=0$ and $\tilde{x} \tilde{y}=1$. Again, these are irreducible sets. On the intersection of the covering sets, we can use the transition maps to find out which of these irreducible subsets are in the same irreducible component of $\pi^{-1}(0)$. For example, for $t \neq 0$, the point parameterized by $\pi_{x}$ as $(x, \tilde{y}, \tilde{z})=\left(0, t^{2}, t\right)$ is the point $\left(p,\left[1: t^{2}: t\right]\right)$ in $X \times \mathbb{P}^{2}$, which is the same point as $\left(p,\left[\frac{1}{t^{2}}: 1: \frac{1}{t}\right]\right)$ parameterized by $\pi_{y}$ as $(\tilde{x}, y, \tilde{z})=\left(\frac{1}{t^{2}}, 0, \frac{1}{t}\right)$, which is the same point as $\left(p,\left[\frac{1}{t}: t: 1\right]\right)$ parameterized by $\pi_{z}$ as $(\tilde{x}, \tilde{y}, z)=\left(\frac{1}{t}, t, 0\right)$. Since these sets are all dense in their respective $\pi_{i}^{-1}(p)$, we can conclude that they form a single irreducible component of the exceptional locus.
$\mathbf{C}_{3}$ : The variety $\mathbb{A}^{2} / / G$ is isomorphic to $V\left(x y-z^{3}\right)$. The blowup is covered by nonsingular affine varieties $\tilde{y}-\tilde{z}^{3} x$ (for $U_{x}$ ), $\tilde{x}-\tilde{z}^{3} y$ (for $U_{y}$ ) and $\tilde{x} \tilde{y}-z$ (for $U_{z}$ ). The respective parametrizations of the subsets of exceptional locus are $V(x, \tilde{y})$ (for $U_{x}$ ), $V(y, \tilde{x})$ (for $U_{y}$ ) and $V(z, \tilde{x} \tilde{y})$. Notice that the first two affine algebraic subsets are irreducible, but the third one decomposes as $V(z, \tilde{x}) \cup V(z, \tilde{y})$. After applying the maps $\pi_{i}$, we can see that the exceptional locus is the set of points $\left\{\left(p,\left[x^{\prime}: y^{\prime}: z^{\prime}\right]\right) \mid x^{\prime} y^{\prime}=0\right\}$, which decomposes into two irreducible components $\left\{\left(p,\left[x^{\prime}: y^{\prime}: z^{\prime}\right]\right) \mid x^{\prime}=0\right\} \cup\left\{\left(p,\left[x^{\prime}: y^{\prime}: z^{\prime}\right]\right) \mid y^{\prime}=0\right\}$. We can see that these two components intersect in one point $(p,[0: 0: 1])$.
$\mathbf{C}_{\mathbf{n}}$ : We will firstly prove by induction on $n$ that the singularities of $X=\mathbb{A}^{2} / / G$ can be resolved via blowups. For $n=2$ and $n=3$ the statement holds, assume $n \geq 4$. The variety $\mathbb{A}^{2} / / G$ is isomorphic to $V\left(x y-z^{n}\right)$. The blowup is covered by nonsingular affine varieties $\tilde{y}-\tilde{z}^{3} x^{n-2}$ (for $U_{x}$ ), $\tilde{x}-\tilde{z}^{3} y^{n-2}$ (for $U_{y}$ ) and a variety $\tilde{x} \tilde{y}-z^{n-2}$ (for $U_{z}$ ), which is singular at $(0,0,0)$. Analogically to $C_{3}$, the exceptional locus decomposes as

$$
\left\{\left(p,\left[x^{\prime}: y^{\prime}: z^{\prime}\right]\right) \mid x^{\prime} y^{\prime}=0\right\}=\left\{\left(p,\left[x^{\prime}: y^{\prime}: z^{\prime}\right]\right) \mid x^{\prime}=0\right\} \cup\left\{\left(p,\left[x^{\prime}: y^{\prime}: z^{\prime}\right]\right) \mid y^{\prime}=0\right\}
$$

with the intersection point $q=(p,[0: 0: 1])$. However, the point $q$ is singular, therefore $B l_{p}(X)$ is not a resolution of singularities. Therefore we need to construct another blowup $B l_{q}\left(B l_{p}(X)\right)$. Luckily for us, in the affine cover of $B l_{p}(X)$, the singularity is covered by the affine variety $\mathbb{A}^{2} / / C_{n-2}$, therefore we already know we can resolve the singularity at $q$ by blowups. Let $f: Y \rightarrow B l_{p}(X)$ be the resolution. Then $Y$ with the composition of morphisms $\pi \circ f$ (where $\pi$ is the blowup morphism $B l_{p}(X) \rightarrow X$ ) is a resolution of singularities of $X$. Firstly, the exceptional locus of this resolution is the set $f^{-1}\left(\left\{\left(p,\left[x^{\prime}: y^{\prime}: z^{\prime}\right]\right) \mid x^{\prime} y^{\prime}=0\right\}\right)$. If we remove the point $q$ from the set $\left\{\left(p,\left[x^{\prime}: y^{\prime}: z^{\prime}\right]\right) \mid x^{\prime} y^{\prime}=0\right\}$, the resulting set lie in $B l_{p}(X)^{\text {reg }}$, therefore is isomorphic to its preimage $f^{-1}\left(\left\{\left(p,\left[x^{\prime}: y^{\prime}: z^{\prime}\right]\right) \mid x^{\prime} y^{\prime}=0\right\} \backslash\{q\}\right)$. Denote $l_{x}=\left\{\left(p,\left[x^{\prime}: y^{\prime}: z^{\prime}\right]\right) \mid x^{\prime}=0\right\}, l_{x}=\left\{\left(p,\left[x^{\prime}: y^{\prime}: z^{\prime}\right]\right) \mid y^{\prime}=0\right\}$. Two irreducible components of the exceptional locus are therefore the closures of $f^{-1}\left(l_{x} \backslash\{q\}\right)$ and $f^{-1}\left(l_{y} \backslash\{q\}\right)$ in $Y$.

We will now prove by induction on $n$ that the Theorem 3.3.9 holds for $G=$ $C_{n}$ and that the components $\overline{f^{-1}\left(l_{x} \backslash\{q\}\right)}, \overline{f^{-1}\left(l_{y} \backslash\{q\}\right)}$ correspond to the leftmost and right-most vertex of the Dynkin diagram. If we want to prove this, we already know from the induction hypothesis how does the exceptional locus of $Y \rightarrow Q$ look like, therefore we only need to show that $\overline{f^{-1}\left(l_{x} \backslash\{q\}\right)}$ and $\overline{f^{-1}\left(l_{y} \backslash\{q\}\right)}$ intersect this locus the way they should. Since the resolution $Y \rightarrow$ $B l_{p}(X)$ is an isomorphism outside $f^{-1}(q)$, we only need to resolve the affine neighborhood $V\left(x y-z^{n-2}\right)$ of $q$ (we have renamed the variables from $\tilde{x}, \tilde{y}, z$ back to $x, y, z$, so we can use the tilded letters for the variables of the second blowup). Therefore if we denote $\tilde{\pi}$ the morphism of the blowup $B l_{q}(V(x y-$ $\underline{\left.z^{n-2}\right)}$ ), we are interested in $\overline{\tilde{\pi}^{-1}\left(l_{x} \backslash\{q\}\right)}=\overline{\{(0, y, 0,[1,0,0])\}}, \overline{\tilde{\pi}^{-1}\left(l_{y} \backslash\{q\}\right)}=$ $\{(x, 0,0,[1,0,0])\}$. The exceptional locus of this blowup contains one irreducible component $\left\{\left(q,\left[x^{\prime}, y^{\prime}, z^{\prime}\right]\right) \in V\left(x y-z^{n-2}\right) \times \mathbb{P}^{2} \mid x^{\prime} y^{\prime}=z^{\prime 2}\right\}$ if $n=4$, two irreducible components $\left\{\left(q,\left[0, y^{\prime}, z^{\prime}\right]\right)\right\}$ and $\left\{\left(q,\left[x^{\prime}, 0, z^{\prime}\right]\right)\right\}$ if $n>4$. In either case, we see that $\overline{\tilde{\pi}^{-1}\left(l_{x} \backslash\{q\}\right)}$ and $\overline{\tilde{\pi}^{-1}\left(l_{y} \backslash\{q\}\right)}$ intersect the components of the locus as they should. Since $\overline{\tilde{\pi}^{-1}\left(l_{x} \backslash\{q\}\right)}$ and $\overline{\tilde{\pi}^{-1}\left(l_{y} \backslash\{q\}\right)}$ doesn't contain any singular points, this fact is not changed by consequent blowups, therefore the statement is proven.

The resolution for the rest of the groups can be found in Hemelsoet 2018]

## Conclusion

We have classified finite subgroups of $\operatorname{SL}(2, C)$ using their maximal abelian subgroups. This gave us some restrictions on the subgroup, we have then one by one checked the possible types of the subgroup and shown whether the subgroup exists and its uniqueness up to conjugation. After the classification of the Dynkin diagrams and extended Dynkin diagrams, we have continued by showing both parts of the classical McKay correspondence. We have shown that the McKay graph of a subgroup of $\mathrm{SL}(2, \mathrm{C})$ is an extended Dynkin diagram and it contains enough information to uniquely determine the group. We also gave an outline of the geometric correspondence by computing GIT-quotients of the two-dimensional complex space and then blowing up the singular points.

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