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# The 2-Kronecker Quiver and Systems of Linear Differential Equations 

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## Problem Description

Let $Q$ be a quiver with two vertices and two arrows going in the same direction, known as the 2-Kronecker quiver. Let $V=\left(V_{1}, V_{2}, A, B\right)$ be a representation of the quiver, where $V_{1}$ and $V_{2}$ are vector spaces over a field $k$, and $A$ and $B$ are linear transformations from $V_{1}$ into $V_{2}$. It can be shown that there are only three classes of indecomposable representations over this quiver. This thesis considers two problems:

1. The main problem of this thesis is the problem of classifying all the indecomposable representations of the 2-Kronecker quiver over an algebraically closed field.
2. The other problem mentioned in this thesis is the problem of solving a system of linear differential equations, $A x=B \dot{x}$, where $A$ and $B$ are $m \times n$-matrices.

## Abstract

We present a way of classifying all the indecomposable representations of the 2-Kronecker quiver over an algebraically closed field. We do this by constructing classes of irregular indecomposable representations by the coxeter functor, and by constructing a class of regular indecomposable representations by mathematical induction using projective resolutions and the Ext ${ }^{1}(\mathrm{~A}, \mathrm{~B})$ functor.

When the indecomposable representations have been classified, we use the decomposition of any representation into a finite direct sum of indecomposable representations to evaluate some systems of linear differential equations on the form $A x=B x^{\prime}$, where $A$ and $B$ are $m \times n$-matrices.

## Sammendrag

I denne oppgaven presenteres en måte å klassifisere de ikke-dekomponerbare representasjonene av 2-Kronecker koggeret (Engelsk:quiver) over en algebraisk lukket kropp. To klasser av ikke-dekomponerbare representasjoner blir bestemt ved å benytte Coxeterfunktoren på noen få, velkjente, ikke-dekomponerbare representasjoner. Den siste klassen blir bestemt ved hjelp av Ext ${ }^{1}$-funktoren og betraktninger om den projektive oppløsningen av ikke-dekomponerbare representasjoner.

I tillegg til å klassifisere de ikke-dekomponerbare representasjonene som beskrevet, blir denne klassifiseringen brukt til å studere lineære likningssystemer av formen $A x=B x^{\prime}$, hvor $A$ og $B$ er $m \times n$-matriser.

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## Introduction

The objective of this thesis is to classify all the indecomposable representations of the 2-Kronecker quiver over an algebraically closed field. We also consider how this classification might prove useful for solving systems of differential equations.
The 2-Kronecker quiver is a quiver with two vertices, and two arrows both going in the same direction.
Given a representation $V$ of a quiver over a field, $k . V$ is indecomposable if it cannot be written as a direct sum of representations, $V=V_{1} \oplus V_{2}$. The focus of this thesis is to classify all indecomposable representations for the given quiver, that is, to find a finite number of classes of representations such that every indecomposable representation of the quiver is an element in one of these classes. The classification of the indecomposable representations over the 2-Kronecker quiver has been known for a long time. In [2], the indecomposable representations have been classified by the use of representation theory for hereditary algebras, and by tools such as the Auslander-Reiten quiver. However, the aim of this thesis is to achieve the classification in a simpler way, by a more ad hoc theoretical approach.
In chapter 1, we establish some basic definitions and properties of modules and sequences of modules, and finally study the baer sum, which is a useful tool in the evaluation of exact sequences.
Chapter 2 is considered to be the main body of this thesis, and contains the construction of the three different classes of indecomposable representations, and a proof that all indecomposable representations are contained in one of these classes.
In chapter 3, we consider one possible application for the classification obtained in chapter 2

For the proofs provided in this thesis, the symbol $\square$ is used to indicate completion of the proof.

## Chapter 1

## Preliminary Results

In this chapter, we will give some basic definitions, and use these to derive some useful properties of modules and sequences of modules. Although the reader is assumed to be familiar with most of the concepts contained in this chapter, the chapter provides a general introduction to some of this theory to make the remainder of the thesis more accessible to any reader. For more about the fundamental background, and for some of the definitions omitted, see [1], [2], [3], and 5].

### 1.1 Free, projective, and injective modules

## Definition.

For a ring, $R$, a left $R$-module, $M$ is an additive abelian group, and a mapping $(r, m) \mapsto r m$ of $R \times M$ into $M$ such that the following holds:

$$
\begin{equation*}
r\left(m_{1}+m_{2}\right)=r m_{1}+r m_{2}, \tag{i}
\end{equation*}
$$

(ii)
$\left(r_{1}+r_{2}\right) m=r_{1} m+r_{2} m$,
(iii) $\left(r_{1} r_{2}\right) m=r_{1}\left(r_{2} m\right)$,
(iv) $1 m=m$, if $1 \in R$
where $r, r_{1}, r_{2} \in R, m, m_{1}, m_{2} \in M$.
Throughout this chapter, unless otherwise stated, whenever we refer to "modules", we really mean left $R$-modules for a ring $R$.

## Definition.

Let $R$ be a ring. Let $A$ and $B$ be left $R$-modules. Then a mapping $f: A \rightarrow B$ is called an $R$-homomorphism if

$$
\begin{array}{ll}
\text { (i) } & f(x+y)=f(x)+f(y) \\
\text { (ii) } & f(r x)=r f(x)
\end{array}
$$

for all $x, y \in A$, and $r \in R$.

## Definition.

Let $R$ be a ring with unity. An $R$-module, $F$, is a free $R$-module if it admits a basis. That is, if there exists a set $X=\left\{x_{j}\right\}_{j \in J} \subseteq F$ such that $X$ is linearly independent, and for each $f \in F$,

$$
f=\sum_{j \in J} c_{j} x_{j}, c_{j} \in R,
$$

and only finitely many $c_{j} \neq 0$.
Proposition 1.1. Let $R$ be a ring with unity, and let $M$ be a left $R$-module, then there exists a free left $R$-module $F$, and a surjective $R$-homomorphism $\phi: F \rightarrow M$.

Proof: One may construct a free left $R$-module, $F$, and an $R$-homomorphism $\phi: F \rightarrow M$ as follows:
Let $F=\left\{h: M \rightarrow R| | M \backslash h^{-1}(0) \mid<\infty\right\}$. (All functions $h$ such that there is a finite number of elements not mapped to zero.)
Let $f_{1}, f_{2} \in F$. Define an addition on $F$ as follows: $\left(f_{1}+f_{2}\right)(m)=f_{1}(m)+_{R}$ $f_{2}(m)$. Here, $+_{R}$ is the addition operator in $R$. By construction, $F$ contains additive inverses and a zero element.
For $r \in R,(r \cdot f)(m)=r \cdot f(m) \in R$. Thus, $F$ is an additive abelian group, and it is easy to confirm that it satisfies the remaining conditions of an $R$-module. To show that $F$ is a free $R$-module, it is enough to show that it is an $R$-module that admits a basis.
To see that $F$ has a basis, consider for each $m \in M$ the kronecker function

$$
\delta_{m}(x)= \begin{cases}1 & \text { if } x=m \\ 0 & \text { if } x \neq m\end{cases}
$$

Now for any $f \in F$,

$$
f(x)=\sum_{m \in M} f(m) \delta_{m}(x),
$$

so the set $\left\{\delta_{m} \mid m \in M\right\}$ spans $F$. Also, this set is linearly independent, since for each $x \in M$, we have:

$$
\sum_{m \in M} r_{m} \delta_{m}(x)=r_{x}, \text { when } r_{i} \in R, \quad \quad \sum_{m \in M} r_{m} \delta_{m}=0 \Rightarrow r_{m}=0 \forall m \in M
$$

Thus, $F$ admits a basis, and $F$ is a free left $R$-module. Define the $R$-homomorphism

$$
\begin{aligned}
\phi: F \rightarrow M, \phi(f) & =\sum_{m \in M} f(m) \cdot m \\
\phi\left(f_{1}+f_{2}\right) & =\phi\left(f_{1}\right)+\phi\left(f_{2}\right) \\
\phi\left(r \cdot f_{1}\right) & =r \cdot \phi\left(f_{1}\right)
\end{aligned}
$$

As $\phi\left(\delta_{m}\right)=m$, this is surjective.
Definition. Let $R$ be a ring. Let $A$ and $B$ be left $R$-modules. A left $R$-module, $P$, is a projective module if for every surjective $R$-homomorphism $f: A \rightarrow B$, and every $R$-homomorphism $g: P \rightarrow B$, there exists an $R$-homomorphism $h: P \rightarrow A$ such that $f h=g$.


Proposition 1.2. A free left $R$-module is projective.

Proof: Let $F$ be a free left $R$-module. Then there exists a set $\mathcal{B}$ s.t. $\mathcal{B}$ is a basis for $F$. Consider any left $R$-modules $A$ and $B$ and any surjective $R$ homomorphism $f: A \rightarrow B$ and any $R$-homomorphism $g: F \rightarrow B$. As $f$ is surjective, one may choose a mapping $h^{\prime}: \mathcal{B} \rightarrow A$ such that

$$
h^{\prime}\left(x^{\prime}\right) \in\left\{a \in A \mid f(a)=g\left(x^{\prime}\right)\right\}, \forall x^{\prime} \in \mathcal{B}
$$

by the axiom of choice. As any element $x \in F$ is uniquely determined by

$$
x=\sum_{x_{i}^{\prime} \in \mathcal{B}} r_{i} x_{i}^{\prime}, r_{i} \in R
$$

this gives rise to an $R$-homomorphism

$$
\begin{aligned}
& h: F \rightarrow A \\
& h(x)=\sum_{x_{i}^{\prime} \in \mathcal{B}} r_{i} h^{\prime}\left(x_{i}^{\prime}\right)
\end{aligned}
$$

which is uniquely determined by the choice of $h^{\prime}$, and we have that $g h(x)=f(x)$. Thus, $F$ is a projective module.


Proposition 1.3. A projective module is a summand of a free module.

Proof: By Proposition 1.1, for any projective module $P$, there exists a free module $F$ and a surjective $R$-homomorphism $\phi: F \rightarrow P$, such that the following diagram commutes:


Now, $\phi \circ h=1_{P} \Rightarrow P \simeq \operatorname{Im} h . F=\operatorname{Im} h \oplus \operatorname{ker} \phi$. Hence, $P$ is a summand of a free module.

Proposition 1.4. Let $P$ be a direct summand of a free $R$-module $F$. Then $P$ is a projective module.

Proof: Let $P$ be a direct summand of a free $R$-module, $F$. Let $h: P \hookrightarrow F$ be an inclusion, and $h^{\prime}: F \rightarrow P$ be an $R$-homomorphism such that $h^{\prime} \circ h=1_{P}$. Let $A$ and $B$ be $R$-modules, such that there exists an $R$-homomorphism $g: P \rightarrow B$, and a surjective $R$-homomorphism $f: A \rightarrow B$. As $F$ is projective by Proposition 1.2 and as $f \circ h^{\prime}: F \rightarrow B$ defines an $R$-homomorphism from $F$ to $B$, there exists an $R$-homomorphism $f^{\prime}: F \rightarrow A$, such that $f \circ f^{\prime}=g \circ h^{\prime}$. Then, by composition by $h$ on the right, one obtains the relation $f \circ f^{\prime} \circ h=g$, hence one have obtained an $R$-homomorphism $f^{\prime} \circ h: P \rightarrow A$, and thus, $P$ is projective.


Proposition 1.5. Let $P_{1}, P_{2}, \ldots, P_{i}$ be projective modules. Then $P=\bigoplus_{j=1}^{i} P_{j}$ is projective.

Proof: Let $P_{1}, P_{2}, \ldots, P_{i}$ be projective modules. By proposition 1.3 , projective modules are summands of free modules. Let $F_{1}, F_{2}, \ldots, F_{i}$ be free modules such that $P_{j}$ is a summand of $F_{j}$ for all $j \in\{1, \ldots, i\}$. Now we have that

$$
P=\bigoplus_{j=1}^{i} P_{j}
$$

is a direct summand of the module

$$
F=\bigoplus_{j=1}^{i} F_{j}
$$

$F$ is a free module by the definition of a free module. Thus, a direct sum of projective modules is a summand of a free module. By proposition 1.4, a summand of a free module is projective.

Definition. Let $R$ be a ring. Let $A$ and $B$ be left $R$-modules. A left $R$-module, $I$ is an injective module if for every injective $R$-homomorphism $f: A \rightarrow B$, and every $R$-homomorphism $g: A \rightarrow I$, there exists an $R$-homomorphism $h: B \rightarrow I$ such that $h f=g$.

$$
\begin{aligned}
& A \xrightarrow{A} B \\
& \underset{I}{g L^{\prime} \exists h}
\end{aligned}
$$

### 1.2 Exact sequences

Definition. An exact sequence of $R$-modules, is a sequence of $R$-modules, $\left\{A_{i}\right\}_{i \in \mathbb{Z}}$, and morphisms $f_{i}: A_{i} \rightarrow A_{i+1}$, such that $\operatorname{Im} f_{i}=\operatorname{ker} f_{i+1}, \forall i \in \mathbb{Z}$,

$$
\cdots \rightarrow A_{i} \xrightarrow{f_{i}} A_{i+1} \xrightarrow{f_{i+1}} A_{i+2} \rightarrow \cdots
$$

Definition. The length of an $R$-module $M$, is defined as $\infty$ or the number $n$ of submodules in the longest chain of submodules of $M$ such that:

$$
N_{0} \subsetneq N_{1} \subsetneq \cdots \subsetneq N_{n}=M,
$$

where $N_{i}$ are submodules of $M$ for $i \in\{1,2, \ldots, n\}$.
Remark. By the Jordan-Hölder Theorem, see [2, Theorem 1.2, p. 3], the length of an $R$-module of finite length, $n$, is independent of the choice of submodules. Remark. For a vector space, $V, \ell(V)=\operatorname{dim}(V)$.

Proposition 1.6. Let

$$
0 \rightarrow A \xrightarrow{f} B \xrightarrow{g} C \xrightarrow{h} D \rightarrow 0
$$

be an exact sequence, where $f$ is a monomorphism, and $h$ is surjective. If $A$, $B, C$, and $D$ are $R$-modules of finite length, then $\ell(A)+\ell(C)=\ell(B)+\ell(D)$.

Proof: Construct the short exact sequence

$$
0 \rightarrow A \xrightarrow{f} B \xrightarrow{p} B / \operatorname{Im} f=E \rightarrow 0,
$$

where $p$ is the projection in the obvious way. That this sequence is "short", means that $f$ is injective and $p$ is surjective. By a corollary of the JordanHölder theorem, see[2, Corollary 1.3, p. 4], one obtains

$$
\begin{equation*}
\ell(A)+\ell(E)=\ell(B) \tag{1.1}
\end{equation*}
$$

Also, there is the short exact sequence

$$
0 \rightarrow E \xrightarrow{g^{\prime}} C \xrightarrow{h} D \rightarrow 0,
$$

where $g^{\prime}: E \rightarrow C$, by $g^{\prime}(b+\operatorname{Im} f)=g(b)$, and hence

$$
\begin{equation*}
\ell(E)+\ell(D)=\ell(C) . \tag{1.2}
\end{equation*}
$$

Combining equation 1.1 and equation 1.2 one obtains equation 1.3 .

$$
\begin{equation*}
\ell(A)+\ell(E)-\ell(D)-\ell(E)=\ell(B)-\ell(C) \Rightarrow \ell(A)+\ell(C)=\ell(B)+\ell(D) \tag{1.3}
\end{equation*}
$$

Definition. An $R$-module $M$ is called noetherian if for every ascending sequence of $R$-submodules of $M$,

$$
M_{1} \subset M_{2} \subset M_{3} \subset \cdots
$$

there exists a positive integer $k$ such that $M_{k}=M_{k+1}=M_{k+2} \cdots$.
Definition. An $R$-module $M$ is called artinian if for every descending sequence of $R$-submodules of $M$,

$$
M_{1} \supset M_{2} \supset M_{3} \supset \cdots
$$

there exists a positive integer $k$ such that $M_{k}=M_{k+1}=M_{k+2} \cdots$.
Remark. An $R$-module of finite length is both artinian and noetherian.

### 1.3 Extension modules

## Definition.

An extension of a module $B$ by $A$, is a short exact sequence

$$
0 \rightarrow B \xrightarrow{f} C \xrightarrow{g} A \rightarrow 0
$$

Definition. A short exact sequence

$$
0 \rightarrow A \xrightarrow{f} B \xrightarrow{g} C \rightarrow 0
$$

is split if there exists a map $j: C \rightarrow B$ with $g j=1_{C}$.
Definition. Two extensions $\alpha: 0 \rightarrow B \xrightarrow{f} C \xrightarrow{g} A \rightarrow 0$ and
$\beta: 0 \rightarrow B \xrightarrow{f^{\prime}} C^{\prime} \xrightarrow{g^{\prime}} A \rightarrow 0$ are equivalent if there exists a map $\phi: C \rightarrow C^{\prime}$ such that the following diagram commutes:


Moreover, by the Five Lemma, see [5, p. 90], if the two extensions $\alpha$ and $\beta$ are equivalent, then $\phi$ is an isomorphism.

Example. $R=\mathbb{Z}$. Take the extensions as follows:

$$
\begin{aligned}
& \alpha: 0 \rightarrow \mathbb{Z}_{3} \xrightarrow{\lambda_{3}} \mathbb{Z}_{9} \xrightarrow{p} \mathbb{Z}_{3} \rightarrow 0 \\
& \beta: 0 \rightarrow \mathbb{Z}_{3} \xrightarrow{\lambda_{6}} \mathbb{Z}_{9} \xrightarrow{p} \mathbb{Z}_{3} \rightarrow 0
\end{aligned}
$$

where $p$ is the projection onto $\mathbb{Z}_{3}$ by isomorphism with the quotient ring $\mathbb{Z}_{9} / \mathbb{Z}_{3}$, and

$$
\lambda_{i}: \mathbb{Z}_{i} \rightarrow \mathbb{Z}_{j}, \lambda_{i}(x)=i x .
$$

To show that $\alpha$ and $\beta$ are not equivalent, it is enough to show that no $\phi: \mathbb{Z}_{9} \rightarrow \mathbb{Z}_{9}$ can make both of the following diagrams commute:


Right side: | ${\underset{\mathbb{Z}}{9}}^{\mathbb{Z}_{9}} \xrightarrow{p} \mathbb{Z}_{3}$ |
| :--- |
| ।। |
| $\mathbb{Z}_{3}$ |

In order to make the left side commute, $\phi(0)=0, \phi(3)=6, \phi(6)=3$. Hence, $\phi=\lambda_{2}$ or $\phi=\lambda_{5}$. But considering the right side, $p \phi(1)=\overline{2} \neq \overline{1}=p(1)$. Hence, the right side diagram does neither commute for $\phi=\lambda_{2}$ nor $\phi=\lambda_{5}$.

## Definition.

Let $X$ be a left $R$-module. Then a projective presentation of $X$ is an exact sequence

$$
P_{1} \xrightarrow{f_{1}} P_{0} \xrightarrow{f_{0}} X \rightarrow 0
$$

where $P_{1}$ and $P_{0}$ are projective modules.

### 1.4 Categories and functors

Definition. A category, $\mathcal{C}$ consists of

1. a class of objects, obj $\mathcal{C}$,
2. a class of morphisms between objects, hom $\mathcal{C}$,
3. and a composition of morphims. This is a binary operation such that when $a, b, c \in \operatorname{obj} \mathcal{C}, \operatorname{hom}(a, b) \times \operatorname{hom}(b, c) \mapsto \operatorname{hom}(a, c)$. Let $f \in \operatorname{hom}(a, b)$, $g \in \operatorname{hom}(b, c)$, then the composition is denoted by $g f \in \operatorname{hom}(a, c)$.
such that
a) the composition of morphisms is associative, $f \in \operatorname{hom}(a, b), g \in \operatorname{hom}(b, c), h \in \operatorname{hom}(c, d),(h g) f=h(g f)$.
b) for every object $x \in \operatorname{obj} \mathcal{C}$, there exists an identity morphism, $1_{x}: x \rightarrow x$, such that for any morphism $f: a \rightarrow b, 1_{b} f=f 1_{a}$.

Definition. Let $\mathcal{C}$ and $\mathcal{D}$ be categories. A covariant functor, $F$ from $\mathcal{C}$ to $\mathcal{D}$ is a mapping that

1. associates to each object $a \in \mathcal{C}$ an object $F(a) \in \mathcal{D}$.
2. associates to each morphism $f: a \rightarrow b \in \mathcal{C}$ a morphism $F(f): F(a) \rightarrow F(b) \in \mathcal{D}$ such that
a) identity morphims are preserved, and
b) compositions are well behaved.

That is, $F\left(1_{a}\right)=1_{F(a)}$, and $F(g f)=F(g) F(f)$ for all objects $a, b, c \in \mathcal{C}$ and all morphims $f, g \in \mathcal{C}$, such that: $f: a \rightarrow b$ and $g: b \rightarrow c$.
Definition. Let $\mathcal{C}$ and $\mathcal{D}$ be categories. A contravariant functor, $F$ from $\mathcal{C}$ to $\mathcal{D}$ is a mapping that

1. associates to each object $a \in \mathcal{C}$ an object $F(a) \in \mathcal{D}$.
2. associates to each morphism $f: a \rightarrow b \in \mathcal{C}$ a morphism
$F(f): F(b) \rightarrow F(a) \in \mathcal{D}$ such that
a) identity morphims are preserved, and
b) compositions are well behaved.

That is, $F\left(1_{a}\right)=1_{F(a)}$, and $F(g f)=F(f) F(g)$ for all objects $a, b, c \in \mathcal{C}$ and all morphims $f, g \in \mathcal{C}$, such that: $f: a \rightarrow b$, and $g: b \rightarrow c$.
Example. We may define a category whose objects are sets, $x$, where the morphisms from the set $x_{1}$ to the set $x_{2}$ are taken to be all mappings of sets from $x_{1}$ to $x_{2}$. This category is called the category of sets, and is denoted by Set.

Example. We may define a category whose objects are all abelian groups, $g$, where the morphisms from the abelian groups $g_{1}$ to the abelian group $g_{2}$ are all group homomorphisms from $g_{1}$ to $g_{2}$. This category is called the category of abelian groups, and is denoted by $A b$.

### 1.5 The functors $\operatorname{Hom}(-, x)$ and $\operatorname{Ext}^{1}(-, x)$

Definition. To each $x \in$ Set, the contravariant Hom-functor, $\operatorname{Hom}(-, x): \mathcal{C} \rightarrow$ Set is a functor given by mapping:

1. an object $a \in \mathcal{C}$ to the set of morphisms mapping $a$ to $x, \operatorname{Hom}(a, x)$, and
2. each morphism $f: a \rightarrow b$ to the function

$$
\begin{aligned}
\operatorname{Hom}(f, x): \operatorname{Hom}(b, x) & \rightarrow \operatorname{Hom}(a, x) \\
g & \mapsto g f, \quad \forall g \in \operatorname{Hom}(b, x)
\end{aligned}
$$

The extension functor, $\operatorname{Ext}^{n}(-, x)$, is a useful functor for canonically constructing exact sequences from short sequences, and repairing exactness lost when using the Hom-functor. Also, there are some useful results when it comes to deciding wheter or not a module is either projective or injective by use of $\operatorname{Ext}^{1}(-, x)$. The derivation of the different characteristics and propositions concerning the $\operatorname{Ext}^{1}(-, x)$ functor is beyond the scope of this paper, but a useful result will be cited to prove the main theorem of chapter 2 .

Proposition 1.7. If $E x t_{R}^{1}(C, A)=\{0\}$, then every extension of $A$ by $C$ splits.
Proof: The proof is omitted in this thesis, but a complete proof of this proposition may be found in [5, p. 421].

Obviously, two exact sequences being equivalent by the definition of equivalence in section 1.3 , is an equivalence relation, and it can be seen that $\operatorname{Ext}^{1}(A, B)$ is really a group of residual classes of exact sequences up to this equivalence relation.

Definition. Let $A$ and $B$ be $R$-modules.
$\operatorname{Ext}_{R}^{1}(A, B)=\{0 \rightarrow B \xrightarrow{f} C \xrightarrow{g} A \rightarrow 0\} / \sim$

### 1.6 Pushout and Pullback diagrams

Definition. For two morphisms $f: Z \rightarrow X, g: Z \rightarrow Y$, the pushout of $f$ and $g$ consists of an object, $P$, and two morphisms $i_{X}: X \rightarrow P$ and $i_{Y}: Y \rightarrow P$, s.t. $i_{X} \circ f=i_{Y} \circ g$.


Also, the pushout must be universal with respect to this diagram. That is, for any other $Q, j_{X}, j_{Y}$ such that $j_{X} \circ f=j_{Y} \circ g$, there must be a unique morphism $u: P \rightarrow Q$ such that $j_{Y}=u \circ i_{Y}$ and $j_{X}=u \circ i_{X}$.

To prove that the pushout exists, it is enough to show that at least one module and pair of morphisms exists satisfying these requirements. For all morphisms $f$ and $g$, we may construct an object, $P$, and morphisms $i_{X}$ and $i_{Y}$ such that: $P=$ Coker $\binom{f}{-g}$. We define the image of $\binom{f}{-g}$ by:

$$
I=\{(f(z),-g(z)) \mid z \in Z\} .
$$

The morphism $i_{X}$ is the compositions of the inclusion $l_{X}$ with the projection $p$, given by the definition of the cokernel, such that $i_{X}=p l_{X}$,

$$
\begin{array}{ll}
l_{X}: X \hookrightarrow X \times Y, & l_{X}(z)=(z, 0) \\
p: X \times Y \rightarrow P, & p(x, y)=(x, y)+I
\end{array}
$$

Thus, we see that $i_{X}(z)=(z, 0)+I$.
The morphism $i_{Y}$ is the composition of the inclusion $l_{Y}$ with the projection $p$, such that $i_{Y}=p l_{Y}$,

$$
l_{Y}: Y \hookrightarrow X \times Y, \quad l_{Y}(z)=(0, z)
$$

From this, we see that $i_{Y}(z)=(0, z)+I$.
This construction is a pushout, and thus, the pushout exists.

Definition. For two morphisms $f: X \rightarrow Z, g: Y \rightarrow Z$, the pullback of $f$ and $g$ consists of an object, $P$, and two morphisms $p_{X}: P \rightarrow X$ and $p_{Y}: P \rightarrow Y$, s.t. the following diagram commutes, that is, $f \circ p_{X}=g \circ p_{Y}$ :


Also, $p_{X}$ and $p_{Y}$ must be universal with respect to this property.

To prove that the pullback exists, it is enough to show that at least one object and two morphisms exists satisfying these requirements.
For all morphisms $f$ and $g$, we may construct an object, $P$, and morphisms $p_{X}$ and $p_{Y}$ such that: $P=\operatorname{ker}\binom{f}{-g}$. The morphism $p_{X}$ is the composition of the inclusion $m$ with the projection $q_{X}$, such that $p_{X}=q_{X} m$,

$$
\begin{array}{ll}
m: P \hookrightarrow X \times Y, & m \text { is inclusion of the kernel of }\binom{f}{-g}, \\
q_{X}: X \times Y \rightarrow X, & q_{X}(x, y)=x .
\end{array}
$$

The morphism $p_{Y}$ is the composition of the inclusion $m$ with the projection $q_{Y}$, such that $p_{Y}=q_{Y} m$,

$$
q_{Y}: X \times Y \rightarrow Y, \quad q_{Y}(x, y)=y
$$

This construction is a pullback, and thus, the pullback exists.

### 1.7 Baer sum

The Baer sum of two extensions can be used as an operation in order to make $\operatorname{Ext}_{R}^{1}(A, B)$ an abelian group. It works by applying pushout and pullback to operate on a pair of extensions,

$$
\begin{aligned}
& \alpha: 0 \rightarrow B \xrightarrow{f} C \xrightarrow{g} A \rightarrow 0 \\
& \beta: 0 \rightarrow B \xrightarrow{f^{\prime}} C^{\prime} \xrightarrow{g^{\prime}} A \rightarrow 0
\end{aligned}
$$

as follows:
Pullback:


By applying "pullback along $A \oplus A$ " one obtains the pullback, $H, j, k$. As the identity and $\binom{1}{1}$ are both injective, $j$ is injective. By commutativity of the right square of the exact sequences, we get:

$$
\begin{aligned}
H & =\left(\begin{array}{ll}
g & 0 \\
0 & g^{\prime}
\end{array}\right)^{-1}\{(a, a) \mid a \in A\} \\
& =\left\{\left(c, c^{\prime}\right) \in C \oplus C^{\prime} \mid g(c)=g^{\prime}\left(c^{\prime}\right)\right\}
\end{aligned}
$$

For the diagram to commute, it needs to satisfy

$$
\binom{1}{1} k\left(c, c^{\prime}\right)=\left(g(c), g^{\prime}\left(c^{\prime}\right)\right),\left(c, c^{\prime}\right) \in H
$$

Thus, we solve the different commutative diagrams and obtain:

$$
\begin{array}{r}
k\left(\left(c, c^{\prime}\right)\right)=g(c) \\
j\left(\left(c, c^{\prime}\right)\right)=\left(c, c^{\prime}\right) \\
i\left(\left(b, b^{\prime}\right)\right)=\left(f(b), f^{\prime}\left(b^{\prime}\right)\right)
\end{array}
$$

## Pushout:

By applying "pushout along $B \oplus B$ ", one obtains $K, m, l$ which is a pushout of $i$ and (11). As both (1 1) and the identity are surjective, $m$ is surjective.

$$
K \simeq \operatorname{Coker}\binom{i}{-1-1} \simeq H / I,
$$

where

$$
I=\left\{\left(f(b),-f^{\prime}(b)\right) \mid b \in B\right\}, I \subseteq H \subseteq C \oplus C^{\prime}
$$

## Baer sum:

$$
\alpha+\beta: 0 \rightarrow B \xrightarrow{l} K \xrightarrow{n} A \rightarrow 0
$$

Where the functions $l$ and $n$ are given as follows:
$l: B \rightarrow K$ is the composition $B \xrightarrow[b \mapsto(f(b), 0)]{ } H \xrightarrow[\left(h, h^{\prime}\right) \mapsto\left(h, h^{\prime}\right)+I]{ } K$
$n: K \rightarrow A$ is uniquely induced by $k: H \rightarrow A$, as $k(I)=0$.

$$
\begin{aligned}
& l: B \rightarrow K, l(b)=(f(b), 0)+I=\left(0,-f^{\prime}(b)\right)+I \\
& n: K \rightarrow A, n\left(\left(k, k^{\prime}\right)+I\right)=g(k) .
\end{aligned}
$$

Example. Consider the extensions

$$
\begin{aligned}
& \alpha: 0 \rightarrow \mathbb{Z}_{3} \xrightarrow{\lambda_{3}} \mathbb{Z}_{9} \xrightarrow{p} \mathbb{Z}_{3} \rightarrow 0 \\
& \beta: 0 \rightarrow \mathbb{Z}_{3} \xrightarrow{\lambda_{6}} \mathbb{Z}_{9} \xrightarrow{p} \mathbb{Z}_{3} \rightarrow 0
\end{aligned}
$$

used in the example of section 1.3 .
To compute the Baer sum, $\gamma=\alpha+\beta$, we first construct a submodule $H$ of $\mathbb{Z}_{9} \oplus \mathbb{Z}_{9}$ such that
$H=\left\{\left(z, z^{\prime}\right) \in \mathbb{Z}_{9} \oplus \mathbb{Z}_{9} \mid p(z)=p\left(z^{\prime}\right)\right\}$. Explicitly, this submodule is given by:

$$
\begin{aligned}
H=\{ & (0,0),(0,3),(0,6),(1,1),(1,4),(1,7),(2,2),(2,5),(2,8), \\
& (3,0),(3,3),(3,6),(4,1),(4,4),(4,7),(5,2),(5,5),(5,8), \\
& (6,0),(6,3),(6,6),(7,1),(7,4),(7,7),(8,2),(8,5),(8,8)\} \simeq \mathbb{Z}_{9} \oplus \mathbb{Z}_{3}
\end{aligned}
$$

Secondly, construct the submodule $I=\left\{\left(f(z),-f^{\prime}(z) \mid z \in \mathbb{Z}_{3}\right\}\right.$ $I=\{(0,0),(3,3),(6,6)\} \simeq \mathbb{Z}_{3}$. Thus,

$$
K \simeq H / I=\{\overline{(0,0)}, \overline{(0,3)}, \overline{(0,6)}, \overline{(1,1)}, \overline{(1,4)}, \overline{(1,7)}, \overline{(2,2)}, \overline{(2,5)}, \overline{(2,8)}\}
$$

Here $\overline{\left(z, z^{\prime}\right)}$ denotes the coset $\left(z, z^{\prime}\right)+I$.
By the isomorphism

$$
\begin{array}{lll}
\phi((0,0))=(0,0), & \phi: K \rightarrow \mathbb{Z}_{3} \oplus \mathbb{Z}_{3} & \\
\phi((1,1))=(1,0), & \phi((1,3))=(0,1), & \phi((0,6))=(1,1),
\end{array}
$$

we have that $K \simeq \mathbb{Z}_{3} \oplus \mathbb{Z}_{3}$.

$$
\begin{aligned}
& l: \mathbb{Z}_{3} \rightarrow \mathbb{Z}_{3} \oplus \mathbb{Z}_{3}, l(z)=(0, z) . \\
& n: \mathbb{Z}_{3} \oplus \mathbb{Z}_{3} \rightarrow \mathbb{Z}_{3}, n\left(z, z^{\prime}\right)=z
\end{aligned}
$$

We have the Baer sum:

$$
\alpha+\beta=\gamma: 0 \rightarrow \mathbb{Z}_{3} \xrightarrow{\left(\begin{array}{ll}
(1)
\end{array}\right.} \mathbb{Z}_{3} \oplus \mathbb{Z}_{3} \xrightarrow{\binom{1}{0}} \mathbb{Z}_{3} \rightarrow 0
$$

Remark. $\alpha, \beta$ and $\gamma$ are all the possible elements of $\operatorname{Ext}_{\mathbb{Z}}^{1}\left(\mathbb{Z}_{3}, \mathbb{Z}_{3}\right)$, and these form an abelian group isomorphic to $\mathbb{Z}_{3}$ !

## Chapter 2

## Kronecker Quiver

In this chapter, we study the finite oriented graph known as the 2-Kronecker Quiver, and try to classify all indecomposable representations of this quiver over an algebraically closed field. Some of the more well-known propositions of this chapter are not proven, but rather referred to from the sources where the reader is provided with complete proofs.

### 2.1 Motivation

The 2-Kronecker Quiver, here denoted by Q, is an oriented graph with two vertices, 1 and 2, and two arrows, $\alpha$ and $\beta$, both going in the same direction, as shown in Figure 2.1.


Figure 2.1: The 2-Kronecker Quiver, Q.
By assigning to each vertex a finite dimensional vector space over a field, $k$, a $k$-vector space, and to each arrow a linear transformation between the vector spaces, we obtain a representation of $Q$ over a field $k$. This representation may be written as a four-tuple ( $k^{m}, k^{n}, l_{\alpha}, l_{\beta}$ ), where $k^{m}$ and $k^{n}$ are the vector spaces assigned to vertices 1 and 2 respectively, and $l_{\alpha}$ and $l_{\beta}$ are linear maps as-
signed to arrows $\alpha$ and $\beta$ respectively. A map between two such representations, $\left(k^{m}, k^{n}, l_{\alpha}, l_{\beta}\right)$ and $\left(k^{m^{\prime}}, k^{n^{\prime}}, l_{\alpha}^{\prime}, l_{\beta}^{\prime}\right)$ is a pair of linear maps,

$$
\begin{aligned}
& f_{1}: k^{m} \rightarrow k^{m^{\prime}} \\
& f_{2}: k^{n} \rightarrow k^{n^{\prime}}
\end{aligned}
$$

such that:

$$
\begin{aligned}
& l_{\alpha}^{\prime} f_{1}=f_{2} l_{\alpha} \\
& l_{\beta}^{\prime} f_{1}=f_{2} l_{\beta}
\end{aligned}
$$

This way we get the category of finite dimensional representations of the quiver, $Q$, where the objects are the representations and the morphisms are the maps between representations.
Let $\Lambda$ be the path algebra defined in equation 2.1.

$$
\Lambda=k Q \simeq\left\{\left.\left(\begin{array}{ccc}
a & 0 & 0  \tag{2.1}\\
b & c & 0 \\
d & 0 & c
\end{array}\right) \right\rvert\, a, b, c, d \in k\right\}
$$

Proposition 2.1. The category of representations of $Q$ over $k$, $\operatorname{rep}(Q, k)$ is equivalent as a category to the category of $\Lambda$-modules of finite $k$-dimension, mod $\Lambda$.

Proof: For a proof of this proposition, see [2, p. 57].
Each finite dimensional module is both artinian and noetherian, so by the Krull-Remak-Schmidt Theorem, see [7. Theorem 3.3, p.7], every $\Lambda$-module may be written as a direct sum of indecomposable $\Lambda$-modules. Thus, in order to prove something for a general $\Lambda$-module, it may be enough to show that it holds for the indecomposable $\Lambda$-modules.

### 2.2 Constructing indecomposable $\Lambda$-modules

Let $J=\left(V_{1}, V_{2}, l_{\alpha}, l_{\beta}\right)$ be interpreted as a $\Lambda$-module corresponding to the representation of $Q$ over $k$, where:
$V_{1}=$ vector space assigned to vertex 1
$V_{2}=$ vector space assigned to vertex 2
$l_{\alpha}=$ linear transformation assigned to arrow $\alpha$
$l_{\beta}=$ linear transformation assigned to arrow $\beta$.

For the remainder of this chapter, the vector spaces assigned to the vertices will be denoted by $k^{n}$ where $n \in \mathbb{N}$, unless otherwise stated. As we have an equivalence of categories between $\bmod \Lambda$ and $\operatorname{rep}(Q, k)$, we use the same notation to describe the representations and the $\Lambda$-modules corresponding to them throughout the text, however, whether the notation is referring to a module or a representation is explicitly stated in every case where it might be unclear.

### 2.2.1 Coxeter functors

To study indecomposable objects of a category of representations of a quiver without oriented cycles, we may use a powerful tool named the coxeter functor, introduced by Bernstein, Gel'fand and Ponomarev, see 8]. The general idea, is that the coxeter functor is a functor constructed in such a way that indecomposable objects are either mapped to other indecomposable objects, or to a zero object. Explicitly, for the case we are concerned with, we have the coxeter functors

$$
\begin{aligned}
& C^{+}: \bmod \Lambda \rightarrow \bmod \Lambda \\
& C^{-}: \bmod \Lambda \rightarrow \bmod \Lambda
\end{aligned}
$$

which maps indecomposable $\Lambda$-modules to indecomposable $\Lambda$-modules or the module $0=(0,0,0,0)$.

## Constructing the coxeter functor

Starting out more general, we have a quiver, $Q^{\prime}$, given by equation 2.2 .

$$
\begin{equation*}
Q^{\prime}=\left\{Q_{0}, Q_{1}, h: Q_{1} \rightarrow Q_{0}, t: Q_{1} \rightarrow Q_{0}\right\} \tag{2.2}
\end{equation*}
$$

$Q_{0}:$ Is a finite set, called the vertices in the quiver.
$Q_{1}$ : Is a finite set, called the arrows of the quiver.
$h(q)$ : Is the head of the arrow $q \in Q_{1}$.
$t(q)$ : Is the tail of the arrow $q \in Q_{1}$.
Head and tail of an arrow is defined such that for an arrow

$$
q: i \rightarrow j, h(q)=j, t(q)=i
$$

Definition. A quiver is said to contain oriented cycles if there exists a finite composition of arrows $q=q_{1} q_{2} \ldots q_{n}$ such that $h(q)=t(q)$.

Definition. A representation of $Q^{\prime}$ over a field $k$, is for each vertex $i \in Q_{0}$ a $k$-vector space, $V(i)$, and for each arrow $q: i \rightarrow j \in Q_{1}$ a linear transformation $l_{q}: V(i) \rightarrow V(j)$.

Remark. Let $Q^{\prime}$ be a quiver with $n$ vertices and $m$ arrows. Then a representation $V$ will be denoted by $V=\left(V(1), \ldots, V(n), l_{1}, \ldots, l_{m}\right)$.

Definition. A map between two representations $V$ and $V^{\prime}$ over the same quiver, $Q^{\prime}$, is defined as a set of linear maps

$$
\left\{f_{i}: V(i) \rightarrow V(i)^{\prime}\right\}_{i \in Q_{0}}
$$

such that if $l_{q}$ and $l_{q}^{\prime}$ are linear transformations assigned to the same arrow $q: i \rightarrow j$ in the different representations, we have:

$$
l_{q}^{\prime} f_{i}=f_{j} l_{q}
$$

This is the same as the linear maps $\left\{f_{i}\right\}_{i \in Q_{0}}$ making the diagram

commute for any $i, j \in Q_{0}$, and $q: i \rightarrow j \in Q_{1}$.

Definition. A sink, is a vertex $i \in Q_{0}$ such that $t(q) \neq i, \forall q \in Q_{1}$.
Definition. The partial coxeter functor of a $\operatorname{sink} i, C_{i}^{+}$, is defined by:

$$
C_{i}^{+}: r e p\left(Q^{\prime}, k\right) \rightarrow r e p\left(Q_{i}^{\prime}, k\right)
$$

where $Q_{i}^{\prime}$ is the quiver obtained when reversing the direction of every arrow $q \in Q^{\prime}$ such that $h(q)=i$. That is, replacing each arrow $q: j \rightarrow i \in Q^{\prime}$ by the arrow $q^{\prime}: i \rightarrow j \in Q_{i}^{\prime}$.
a) Let $V(j)$ be the vector space assigned to vertex $j \in Q_{0}$.

$$
C_{i}^{+}(V(j))= \begin{cases}V(j), & j \neq i \\ V(i)^{\prime}=\operatorname{ker} g, & j=i .\end{cases}
$$

where

$$
g: \bigoplus_{\substack{q \in Q_{1} \\ h(q)=i}} V(t(q)) \longrightarrow V(i) .
$$

Let $v_{t(q)} \in V(t(q))$, then $g$ is defined by

$$
g\left(\left(v_{t(q)}\right)_{\substack{q \in Q_{1} \\ h(q)=i}}\right)=\sum_{\substack{q \in Q_{1} \\ h(q)=i}} l_{q}\left(v_{t(q)}\right) .
$$

b) Let $l_{q}: V(j) \rightarrow V(k)$, be the linear transformation assigned to the arrow $q: j \rightarrow k \in Q_{1}$.

$$
C_{i}^{+}\left(l_{q}\right)= \begin{cases}l_{q}, \\ l_{q^{\prime}}: C_{i}^{+}(V(i)) \xrightarrow{\text { incl. }} \bigoplus_{\substack{q \in Q_{1} \\ h(q)=i}} V(t(q)) \xrightarrow{\text { proj. }} V(t(q)), & k \neq i, \\ k=i .\end{cases}
$$

Assuming the quiver we are considering does not have any oriented cycles, we may describe how $C^{+}$acts on a representation of the quiver, $V$, by the following algorithm:

1. Locate a sink, $i$.
2. Apply $C_{i}^{+}$to the representation.
3. Locate a previously unchanged sink, $j \neq i$, in the quiver $Q_{i}^{\prime}$.
4. Apply $C_{j}^{+}$to the representation $C_{i}^{+}(V)$.
5. Repeat the procedure until every vertex in the quiver have been a sink exactly once.

In other words, by changing the numbering of our vertices to fit to the order in which they appear as a sink for our given choice of order of operations, we may write:

$$
C^{+}(V)=C_{n}^{+} C_{n-1}^{+} \ldots C_{2}^{+} C_{1}^{+}(V)
$$

Definition. A source, is a vertex $i \in Q_{0}$ such that $h(q) \neq i, \forall q \in Q_{1}$.
Definition. The partial coxeter functor of a source $i, C_{i}^{-}$, is defined by:

$$
C_{i}^{-}: \operatorname{rep}\left(Q^{\prime}, k\right) \rightarrow \operatorname{rep}\left(Q_{i}^{\prime \prime}, k\right)
$$

where $Q_{i}^{\prime \prime}$ is the quiver obtained when reversing the direction of every arrow $q \in Q^{\prime}$ such that $t(q)=i$.
a) Let $V(j)$ be the vector space assigned to vertex $j \in Q_{0}$.

$$
C_{i}^{-}(V(j))= \begin{cases}V(j), & j \neq i \\ V(i)^{\prime}=\text { Coker } g, & j=i\end{cases}
$$

where

$$
g: V(i) \rightarrow \bigoplus_{\substack{q \in Q_{1} \\ t(q)=i}} V(h(q))
$$

Let $v_{i} \in V(i)$, then $g$ is defined by:

$$
g\left(v_{i}\right)=\bigoplus_{\substack{q \in Q_{1} \\ t(q)=i}} l_{q}\left(v_{i}\right)
$$

b) Let $l_{q}: V(j) \rightarrow V(k)$, be the linear transformation assigned to the arrow $q: j \rightarrow k \in Q_{1}$.

$$
C_{i}^{-}\left(l_{q}\right)= \begin{cases}l_{q}, \\ l_{q^{\prime}}: V(k) \xrightarrow{\text { incl. }} \bigoplus_{\substack{q \in Q_{1} \\ h(q)=i}} V(h(q)) \xrightarrow{\text { proj. }} C_{i}^{-}(V(i)), \quad j=i, \\ & j=i .\end{cases}
$$

Assuming the quiver we are considering does not contain any oriented cycles, we may describe how $C^{-}$acts upon a representation of a quiver, $V$, by the following algorithm:

1. Locate a source, $i$.
2. Apply $C_{i}^{-}$to the representation.
3. Locate a previously unchanged source in $Q_{i}^{\prime \prime}, j \neq i$.
4. Apply $C_{j}^{-}$to the representation $C_{i}^{-}(V)$.
5. Repeat the procedure until every vertex has appeared as a source exactly once.

In other words, by changing the numbering of our vertices to fit to the order in which they appear as a source, we may write:

$$
C^{-}(V)=C_{n}^{-} C_{n-1}^{-} \cdots C_{2}^{-} C_{1}^{-}(V)
$$

Remark. For any indecomposable representation $V \not \not S_{i} . C_{i}^{-} C_{i}^{+}(V)=V . S_{i}$ is the representation with a one dimensional vector space assigned to the vertex $i$, zero spaces assigned to every vertex $j \neq i$, and zero maps assigned to every arrow $q \in Q_{1}$.
Example. Consider the quiver without cycles given by:


Construct a representation of this quiver over a field $k$ by assigning to each vertex $i \in\{1,2,3\}$ a vector space $V(i)$ over $k$, and to each arrow $\chi \in\{\alpha, \beta, \gamma\}$ a linear map $l_{\chi}$. Denote such a representation by

$$
V=\left(V(1), V(2), V(3), l_{\alpha}, l_{\beta}, l_{\gamma}\right) .
$$

Now consider the representation $V_{1}=\left(k, k^{2}, k^{3},\binom{1}{0},\left(\begin{array}{ll}1 & 0 \\ 0 & 1 \\ 0 & 0\end{array}\right),\left(\begin{array}{ll}0 & 0 \\ 1 & 0 \\ 0 & 1\end{array}\right)\right)$.

$$
k \xrightarrow{\binom{1}{0}} k^{2} \xrightarrow[\left(\begin{array}{ll}
0 & 0 \\
1 & 0 \\
0 & 1
\end{array}\right)]{\left(\begin{array}{ll}
1 & 0 \\
0 & 1 \\
0 & 0
\end{array}\right)} k^{3}
$$

Applying the coxeter functor $C^{+}$to this representation will give us $C^{+}\left(V_{1}\right)=$ $C_{1}^{+} C_{2}^{+} C_{3}^{+}\left(V_{1}\right)$. Starting out, there is only one sink to consider: $V(3)$. Computing, we get:

$$
\begin{gathered}
C_{3}^{+}(V(3))=k e r g, g=\left(l_{\beta}, l_{\gamma}\right): k^{2} \oplus k^{2} \rightarrow k^{3}, \\
C_{3}^{+}(V(3))=\operatorname{ker}\left(\begin{array}{ccccc}
1 & 0 & 0 \\
0 & 1 & 1 & 0 \\
0 & 0 & 0 & 1
\end{array}\right)=\left\{\left.\left(\begin{array}{c}
0 \\
a \\
-a \\
0
\end{array}\right) \right\rvert\, a \in k\right\} \simeq k .
\end{gathered}
$$

The new representation, $V_{1}^{\prime}$, obtained is:

$$
k \xrightarrow{\binom{1}{0}} k^{2} \stackrel{\binom{1}{0}}{\stackrel{\binom{0}{1}}{ }} k
$$

This quiver has a new sink, the vertex 2, and hence, we need to compute $C_{2}^{+}\left(V^{\prime}\right)$ in the same way:

$$
C_{2}^{+}(V(2))=\operatorname{ker}\left(\begin{array}{lll}
1 & 1 & 0 \\
0 & 0 & 1
\end{array}\right)=\left\{\left.\left(\begin{array}{c}
a \\
-a \\
0
\end{array}\right) \right\rvert\, a \in k\right\} \simeq k .
$$

The new representation, $V_{1}^{\prime \prime}$, obtained is:


Now, there is only one vertex left to consider, 1, and this is actually turned into a sink from the last use of the partial coxeter functor, so we apply $C_{1}^{+}\left(V^{\prime \prime}\right)$, and finally obtain $C^{+}(V)$ :

$$
\begin{gathered}
C_{1}^{+}(V(1))=\operatorname{ker} 1=0 . \\
0 \xrightarrow{0} k \xrightarrow[0]{\stackrel{1}{\longrightarrow}} k
\end{gathered}
$$

The quiver is now the same quiver we started with, and we see that

$$
C^{+}\left(\left(k, k^{2}, k^{3},\binom{1}{0},\left(\begin{array}{cc}
1 & 0 \\
0 & 1 \\
0 & 0
\end{array}\right),\left(\begin{array}{cc}
0 & 0 \\
1 & 0 \\
0 & 1
\end{array}\right)\right)\right)=(0, k, k, 0,1,0) .
$$

In a similar way, we may apply $C^{-}$to the representation $V_{2}=C^{+}\left(V_{1}\right)$ by considering the sources of the quiver. First, we consider the source at vertex 1:

$$
C_{1}^{-}(V(1))=\text { Coker } 0=k / \operatorname{Im} 0 \simeq k
$$

The new representation, $V_{2}^{\prime}$, obtained is:


Now the source is located at vertex 2:

$$
C_{2}^{-}(V(2))=\operatorname{Coker}\left(\begin{array}{l}
0 \\
1 \\
0
\end{array}\right)=k \oplus k \oplus k / \operatorname{Im}\left(\begin{array}{l}
0 \\
1 \\
0
\end{array}\right) \simeq k^{2}
$$

The new representation, $V_{2}^{\prime \prime}$, obtained is:


And finally, the last source is now at vertex 3:

$$
C_{3}^{-}(V(3))=\operatorname{Coker}\left(\begin{array}{l}
0 \\
0 \\
0 \\
1
\end{array}\right)=k \oplus k \oplus k \oplus k / \operatorname{Im}\left(\begin{array}{l}
0 \\
0 \\
0 \\
1
\end{array}\right) \simeq k^{3}
$$

The new representation obtained is:


And hence, we see that: $C^{-}\left(V_{2}\right)=V_{1}$, and $C^{-} C^{+}\left(V_{1}\right)=V_{1}$.

### 2.2.2 Coxeter functor and indecomposables of $Q$.

Let $V=\left(k^{m}, k^{n}, A, B\right)$ be a representation of the quiver $Q$, as defined in section 2.1. In order to obtain all the indecomposable representations of $Q$ over a field $k$, or "the indecomposables of $Q$ ", one possible idea is to start out by finding different types of indecomposables and obtain indecomposables of a similar form by applying the coxeter functor to these. The indecomposable representations are either irregular, or regular, and the irregular indecomposables may all be derived from a finite number of indecomposable representations simply by applying the coxeter functors $C^{+}$or $C^{-}$.[8, Theorem 1.3, p. 25].
Proposition 2.2. Let $V=\left(k^{m}, k^{n}, A, B\right)$ be a representation of the quiver $Q$ over the field $k$, such that $m>n$. Now $C^{+}(V)$ is indecomposable if and only if $V$ is indecomposable.

Proof: Let $\Lambda$ be the path algebra defined in equation 2.1. A $\Lambda$-module, $M$, of finite length, is indecomposable if and only if the endomorphism ring $\operatorname{End}_{\Lambda}(M)$ is local, see [2, Theorem 2.2, p. 33]. So in order to prove that the coxeter functor preserves indecomposable modules, it is enough to observe the endomorphism rings of $M$ and $C^{+}(M)$. Consider the diagram:


As the diagram commutes, any homomorphism $f_{i}: V(i) \rightarrow V(i)$ induces a homomorphism $g_{i}: C_{i}^{+}(V(i)) \rightarrow C_{i}^{+}(V(i))$, and the converse is also true.
$\Rightarrow)$ Assume that $M$ is indecomposable, and $M \nsucceq S_{i}$. We have that $\operatorname{End}(M) \simeq$ $\operatorname{End}\left(C^{+}(M)\right)$, and as $\operatorname{End}(M)$ is local, $\operatorname{End}\left(C^{+}(M)\right)$ is local, and thus, $C^{+}(M)$ is an indecomposable module.
$\Leftarrow)$ Assume that $C^{+}(M)$ is indecomposable, and $C_{i}^{+}(M) \nsucceq S_{i}$ for the quiver $Q_{i}$. We have that $\operatorname{End}\left(C^{+}(M)\right) \simeq \operatorname{End}(M)$, and as $\operatorname{End}\left(C^{+}(M)\right)$ is local, so is $\operatorname{End}(M)$, and hence, $M$ is an indecomposable module.
A $\Lambda$-module $M$ is indecomposable if and only if $C^{+}(M)$ is indecomposable. The same holds for representations of the quiver $Q$ over $k$, by proposition 2.1

Proposition 2.3. Let $V=\left(k^{m}, k^{n}, A, B\right)$ be a representation of the quiver $Q$ over the field $k$, such that $m<n$. Now $C^{-}(V)$ is indecomposable if and only if $V$ is indecomposable.

Proof: The proof is similar to the proof of proposition 2.2 .

## General requirements for indecomposable modules

Before we start considering the use of the coxeter functor on an indecomposable representation, we first need to derive some basic properties for the indecomposable representations, or by equivalence, for the indecomposable $\Lambda$-modules.

Proposition 2.4. Let $J=\left(k^{m}, k^{n}, A, B\right)$ be an indecomposable $\Lambda$-module, with $m>0$. Then $\operatorname{Im} A+\operatorname{Im} B=k^{n}$

Proof: Assume $\operatorname{Im} A+\operatorname{Im} B=k^{n^{\prime}}$. Let $n^{\prime \prime}=n-n^{\prime}$. One might write

$$
\left.J=\left(k^{m}, k^{n}, A, B\right)=\left(k^{m}, k^{n^{\prime}}, \bar{A}, \bar{B}\right) \oplus\left(0, k^{n^{\prime \prime}}, 0_{n^{\prime \prime}}, 0_{n^{\prime \prime}}\right)\right)
$$

where $0_{n^{\prime \prime}}$ is the zero matrix of dimensions $1 \times n^{\prime \prime}$, and $\bar{A}$ and $\bar{B}$ are $A$ and $B$ when removing the rows simultaneously taking every element to zero for both $A$ and $B . J$ is decomposable for any $n^{\prime \prime} \neq 0$.

Proposition 2.5. Let $J=\left(k^{m}, k^{n}, A, B\right)$ be an indecomposable $\Lambda$-module, with $n>0$. Then ker $A \cap \operatorname{ker} B=0$

Proof: Let $K=\operatorname{ker} A \cap \operatorname{ker} B=k^{m^{\prime}}$. Then

$$
J=\left(k^{m-m^{\prime}}, k^{n}, \bar{A}, \bar{B}\right) \oplus\left(k^{m^{\prime}}, 0,0^{m^{\prime}}, 0^{m^{\prime}}\right)
$$

where $0^{m^{\prime}}$ is the zero matrix of dimensions $m^{\prime} \times 1$, and $\bar{A}$ and $\bar{B}$ are $A$ and $B$ restricted to the vector space $k^{m} \backslash K$. Thus $J$ is decomposable for any $m^{\prime} \neq 0$.

## Coxeter functor of a representation of $Q$.

Let $V=\left(k^{m}, k^{n}, A, B\right)$ be an indecomposable representation of the quiver $Q$ over $k$.


Obviously, for the quiver, $Q$, there is only one sink, the vertex 2 . Hence, we may let the coxeter functor $C^{+}$act on $V$ by:

$$
C^{+}(V)=C_{1}^{+} C_{2}^{+}(V)
$$

Starting out:

$$
C_{2}^{+}\left(k^{n}\right)=\operatorname{ker}(A B)
$$

And we get the new representation, $V^{\prime}$, such that:

$$
V^{\prime}: \quad k^{m \leftrightarrows} A^{\prime} B^{\prime}<\operatorname{ker}(A B)
$$

Here, $\operatorname{dim}(\operatorname{ker}(A B))$ is given by constructing the short exact sequence:

$$
0 \rightarrow \operatorname{ker}(A B) \hookrightarrow k^{m} \oplus k^{m} \xrightarrow{(A B)} k^{n} \rightarrow 0
$$

where $(A B)$ is surjective by proposition 2.4 . Now as the length of a vector space is equal to its dimension, we get that $\operatorname{dim}(\operatorname{ker}(A B))=m+m-n=$ $2 m-n$, and hence, $\operatorname{ker}(A B) \simeq k^{2 m-n}$.

$$
\begin{gathered}
C^{+}(V)=C_{1}^{+}\left(V^{\prime}\right) \\
C_{1}^{+}\left(k^{n}\right) \simeq \operatorname{ker}\left(A^{\prime} B^{\prime}\right),
\end{gathered}
$$

And we get the representation $C^{+}(V)$ such that:

$$
C^{+}(V): \quad k e r\left(A^{\prime} B^{\prime}\right) \xrightarrow[B^{\prime \prime}]{\stackrel{A^{\prime \prime}}{\Longrightarrow}} k^{2 m-n}
$$

By constructing an exact sequence in the similar way as before, we get

$$
0 \rightarrow \operatorname{ker}\left(A^{\prime} B^{\prime}\right) \hookrightarrow k^{2 m-n} \oplus k^{2 m-n} \xrightarrow{\left(A^{\prime} B^{\prime}\right)} k^{n} \rightarrow 0
$$

$\operatorname{dim}\left(\operatorname{ker}\left(A^{\prime} B^{\prime}\right)\right)=2 m-n+2 m-n-m=3 m-2 n, \operatorname{ker}\left(A^{\prime} B^{\prime}\right) \simeq k^{3 m-2 n}$.

$$
C^{+}\left(k^{m}, k^{n}, A, B\right)=\left(k^{3 m-2 n}, k^{2 m-n}, A^{\prime \prime}, B^{\prime \prime}\right)
$$

By this relation, we may define the coxeter matrix, $\Phi$, which is a matrix describing what happens to the dimensions of the vector spaces in an indecomposable $\Lambda$-module when applying the coxeter functor $C^{+}$on it.

### 2.2.3 Coxeter Matrix

Let the dimensions of an indecomposable $\Lambda$-module

$$
J=\left(k^{m}, k^{n}, A, B\right)
$$

be given by the dimension vector

$$
d_{J}=\binom{m}{n} .
$$

When we use the coxeter functor $C^{+}$to move from one indecomposable $\Lambda$ module, $J_{1}$, to another indecomposable $\Lambda$-module, $J_{2}=C^{+}\left(J_{1}\right)$, the dimension vectors $d_{J_{1}}$ and $d_{J_{2}}$ are related by the coxeter matrix, $\Phi$ :

$$
\Phi=\left(\begin{array}{ll}
3 & -2 \\
2 & -1
\end{array}\right)
$$

Applying the coxeter functor $C^{+}$on a module $q$ times, will change the dimension vectors by $\Phi^{q}$, given by:

$$
\Phi^{q}=q \cdot\left(\begin{array}{ll}
2 & -2 \\
2 & -2
\end{array}\right)+\left(\begin{array}{ll}
1 & 0 \\
0 & 1
\end{array}\right)
$$

This is given by the binomial expression of $(A+I)^{q}$ for quadratic matrices $A=\left(\begin{array}{ll}2 & -2 \\ 2 & -2\end{array}\right)$ and $I=\left(\begin{array}{ll}1 & 0 \\ 0 & 1\end{array}\right)$, as $\left(\begin{array}{ll}2 & -2 \\ 2 & -2\end{array}\right)^{2}=\left(\begin{array}{ll}0 & 0 \\ 0 & 0\end{array}\right)$.

The inverse coxeter functor, $C^{-}$may be studied in a way similar to how we approached $C^{+}$, and it gives an inverse coxeter matrix, $\Phi^{-1}$, such that:

$$
\Phi^{-1}=\left(\begin{array}{ll}
-1 & 2 \\
-2 & 3
\end{array}\right)
$$

Applying the inverse coxeter functor $C^{-}$to a module, $V, q$ times, will change the dimension vectors by $\Phi^{-q}$, given by:

$$
\Phi^{-q}=q \cdot\left(\begin{array}{ll}
-2 & 2 \\
-2 & 2
\end{array}\right)+\left(\begin{array}{ll}
1 & 0 \\
0 & 1
\end{array}\right)
$$

We have that

$$
\Phi \Phi^{-1}=\Phi^{-1} \Phi=\left(\begin{array}{ll}
1 & 0 \\
0 & 1
\end{array}\right)
$$

### 2.2.4 Indecomposables in the class $N_{1}$

To find indecomposable $\Lambda$-modules, it is a good idea to start with the simple modules $S_{1}=(k, 0,0,0)$ and $S_{2}=(0, k, 0,0)$, which are obviously indecomposable. By constructing injective $\Lambda$-modules of the simple $\Lambda$-modules $S_{1}$ and $S_{2}$, we obtain the injective $\Lambda$-modules

$$
I_{1} \simeq S_{1}=(k, 0,0,0)
$$

and

$$
I_{2}=\left(k^{2}, k,\left(\begin{array}{ll}
1 & 0
\end{array}\right),\left(\begin{array}{ll}
0 & 1
\end{array}\right)\right)
$$

respectively. Applying the coxeter functor $C^{+}$to $I_{1}$ gives us:

$$
C^{+}\left(I_{1}\right)=C_{1}=\left(k^{3}, k^{2},\left(\begin{array}{ccc}
1 & 0 & 0 \\
0 & 1 & 0
\end{array}\right),\left(\begin{array}{lll}
0 & 1 & 0 \\
0 & 0 & 1
\end{array}\right)\right)
$$

Applying the coxeter matrix $C^{+}$on $I_{2}$, gives us the module:

$$
C_{2}=\left(k^{4}, k^{3},\left(\begin{array}{cccc}
1 & 0 & 0 & 0 \\
0 & 1 & 0 & 0 \\
0 & 0 & 1 & 0
\end{array}\right),\left(\begin{array}{cccc}
0 & 1 & 0 & 0 \\
0 & 0 & 1 & 0 \\
0 & 0 & 0 & 1
\end{array}\right)\right)
$$

Now,

$$
\begin{gathered}
C_{1}=C^{+}\left(I_{1}\right) \\
C_{2}=C^{+}\left(I_{2}\right),
\end{gathered}
$$

and both $C_{1}$ and $C_{2}$ are indecomposable $\Lambda$-modules, as $I_{1}$ and $I_{2}$ are indecomposable $\Lambda$-modules. Expanding on this idea, by applying $C^{+}$consecutively on the modules attained this way, we may construct an entire class of indecomposable $\Lambda$-modules on the form of equation 2.3 ,

$$
\begin{equation*}
N_{1}=\left\{\left(k^{n+1}, k^{n},\left(i_{n} 0\right),\left(0 i_{n}\right)\right) \mid n \in \mathbb{N}\right\} \tag{2.3}
\end{equation*}
$$

Where $i_{n}$ for the rest of this thesis denotes the $n \times n$-identity matrix.

### 2.2.5 Indecomposables in the class $N_{2}$

Construct the projective $\Lambda$-modules

$$
P_{1}=\left(k, k^{2},\binom{1}{0},\binom{0}{1}\right)
$$

and

$$
P_{2} \simeq S_{2}=(0, k, 0,0)
$$

over $S_{1}$ and $S_{2}$ respectively. We obtain another class of indecomposable $\Lambda$ modules on the form given in equation 2.4 by applying $C^{-}$to $P_{1}$ and $P_{2}$ in a similar manner to the construction we performed in section 2.2.4

$$
\begin{equation*}
N_{2}=\left\{\left.\left(k^{n}, k^{n+1},\binom{i_{n}}{0},\binom{0}{i_{n}}\right) \right\rvert\, n \in \mathbb{N}\right\} . \tag{2.4}
\end{equation*}
$$

### 2.2.6 Indecomposables in the class $N_{3}$

## Jordan canonical form

The last class of indecomposable $\Lambda$-modules we construct, may be simplified by restricting our study to deal with the cases where $k$ is an algebraically closed field, that is, all the irreducible polynomials in the polynomial ring $k[x]$ are of degree 1 . When this is the case, any square matrix $A$, being a linear transformation from $k^{n}$ to $k^{n}$, may be written on Jordan canonical form as a direct product of matrices on the form of equation 2.5, called Jordan blocks. For further explaination of the Jordan canonical form of a matrix, see for instance [1, p. 423].

$$
J B_{n_{i}}^{\lambda}=\left(\begin{array}{cccccc}
\lambda & 1 & 0 & \cdots & 0 & 0  \tag{2.5}\\
0 & \lambda & 1 & \cdots & 0 & 0 \\
0 & 0 & \lambda & \cdots & 0 & 0 \\
\vdots & \vdots & \vdots & \ddots & \vdots & \vdots \\
0 & 0 & 0 & \cdots & \lambda \\
0 & 0 & \cdots & \cdots & 0 & \lambda
\end{array}\right)
$$

$J B_{n_{i}}^{\lambda}$ is an $n_{i} \times n_{i}$-matrix, where $\lambda$ is an eigenvalue of the matrix $A$. Let the algebraic multiplicity, $m_{a}^{\lambda}$ of a fixed eigenvalue $\lambda$ of $A$ be given by:

$$
m_{a}^{\lambda}=\sum_{\substack{n_{i} \in \mathbb{N}, \exists J B_{n_{i}}^{\lambda}}} n_{i}
$$

Described with words, $m_{a}^{\lambda}$ is the dimension of a direct sum of all Jordan blocks in the Jordan canonical form of $A$ containing the eigenvalue $\lambda$. The number of different such Jordan blocks, $m_{g}^{\lambda}$, is called the geometric multiplicity of the eigenvalue $\lambda$. The dimension of the direct sum of all Jordan blocks in the Jordan canonical form of a matrix, $A$, is given by

$$
\sum_{\lambda} m_{a}^{\lambda}=n
$$

where the sum is taken over every different eigenvalue of $A$.

## Indecomposables in the class $N_{3}$

Using the well behaved structure of the Jordan blocks, we may construct a third class of indecomposable $\Lambda$-modules by considering the $\Lambda$-module

$$
S_{\lambda}^{*}=(k, k, 1, \lambda), \lambda \in k \cup\{\infty\}
$$

where $\lambda=\infty$ corresponds to the module $(k, k, 0,1)$.
We see that:

$$
\operatorname{dim}\left(\operatorname{Ext}_{k}^{1}\left(S_{\lambda}^{*}, S_{\gamma}^{*}\right)\right) \begin{cases}1 & \text { if } \lambda=\gamma \\ 0 & \text { if } \lambda \neq \gamma\end{cases}
$$

We obtain a class of indecomposable modules by taking $S_{\lambda}^{*}, \lambda \in k$ extended with itself. These are the $\Lambda$-modules in the class $N_{3}^{\prime}$, given by equation 2.6

$$
\begin{equation*}
N_{3}^{\prime}=\left\{\left(k^{n}, k^{n}, i_{n}, J B_{n}^{\lambda}\right) \mid n \in \mathbb{N}, \lambda \in k\right\} \tag{2.6}
\end{equation*}
$$

Also, $S_{\infty}^{*}$ extended with itself gives us the $\Lambda$-modules in the class $N_{3}^{\prime \prime}$, given by equation 2.7 .

$$
\begin{equation*}
N_{3}^{\prime \prime}=\left\{\left(k^{n}, k^{n}, J B_{n}^{0}, i_{n}\right) \mid n \in \mathbb{N}\right\} \tag{2.7}
\end{equation*}
$$

The modules in the class $N_{3}^{\prime \prime}$ may be denoted as modules of the form of equation 2.6 with $\lambda=\infty$. Thus, the classes $N_{3}^{\prime}$ and $N_{3}^{\prime \prime}$ may be written as a single class of $\Lambda$-modules, $N_{3}$, given by equation 2.8 .

$$
\begin{equation*}
N_{3}=\left\{\left(k^{n}, k^{n}, i_{n}, J B_{n}^{\lambda}\right) \mid n \in \mathbb{N}, \lambda \in k \cup\{\infty\}\right\} \tag{2.8}
\end{equation*}
$$

### 2.3 Main Theorem

The three classes $N_{1}, N_{2}$, and $N_{3}$ does in fact cover all the indecomposable $\Lambda$-modules, as will be shown in the next section. Due to this fact, we arrive at the main theorem of this thesis:

Theorem 2.6. Let $k$ be an algebraically closed field. Let $M$ be a $\Lambda$-module. Then:

$$
\begin{aligned}
M & \simeq\left(k^{m_{1}}, k^{m_{1}+1},\binom{i_{m_{1}}}{0},\binom{0}{i_{m_{1}}}\right) \oplus \cdots \oplus\left(k^{m_{r}}, k^{m_{r}+1},\binom{i_{m_{r}}}{0},\binom{0}{i_{m_{r}}}\right) \\
& \oplus\left(k^{n_{1}+1}, k^{n_{1}},\left(i_{n_{1}} 0\right),\left(0 i_{n_{1}}\right)\right) \oplus \cdots \oplus\left(k^{n_{s}+1}, k^{n_{s}},\left(i_{n_{s}} 0\right),\left(0 i_{n_{s}}\right)\right) \\
& \oplus\left(k^{l_{1}}, k^{l_{1}}, i_{l_{1}}, J B_{l_{1}}^{\lambda_{1}}\right) \oplus \cdots \oplus\left(k^{l_{t}}, k^{l_{t}}, i_{l_{t}}, J B_{l_{t}}^{\lambda_{t}}\right), r, s, t \in \mathbb{N}, \\
& \lambda_{t} \in k \cup\{\infty\}
\end{aligned}
$$

Where $i_{n}$ is the identity matrix of dimension $n \times n$, and $J B_{l_{i}}^{\lambda_{i}}$ is the Jordan block of dimension $l_{i} \times l_{i}$ with eigenvalue $\lambda_{i}$.

### 2.4 Proof of Main Theorem

If $N_{1}, N_{2}$ and $N_{3}$ are all indecomposable $\Lambda$-modules, Theorem 2.6 follows immediately from the Krull-Remak-Schmidt theorem, see [7, Theorem 3.3, p.7]. Hence, we need only show that the classes $N_{1}, N_{2}$, and $N_{3}$ contain all indecomposable $\Lambda$-modules. Considering an arbitrary indecomposable $\Lambda$-module

$$
J=\left(k^{m}, k^{n}, A, B\right)
$$

the cases we need to study may be divided into three separate cases:

1. The case where $m<n$.
2. The case where $m>n$.

3 . The case where $m=n$.
These three will turn out to correspond to the three different classes of indecomposable modules discovered in the previous section.

### 2.4.1 Indecomposables in the classes $N_{2}$ and $N_{1}$

In the following proofs, the specifics of the linear transformations in the representations are not always shown, as the proofs generally rely on dimension arguments from using the coxeter functor.

Proposition 2.7. Let $J=\left(k^{m}, k^{n}, A, B\right)$ be an indecomposable $\Lambda$-module. If $m<n$, then $n=m+1$.

Proof: Assuming the module $J$ is not isomorphic to $P_{1}$ or $P_{2}$, in which case it would belong to the class $N_{2}$, we may use the coxeter functor $C^{+}$on the module, to obtain a $\Lambda$-module on the form $\left(k^{3 m-2 n}, k^{2 m-n}, A^{\prime \prime}, B^{\prime \prime}\right)$. By continued application of the coxeter functor, the dimension of the vector spaces will be given by $\Phi^{q}$, as defined in section 2.2.3. This yields modules

$$
\left(k^{(2 q+1) m-2 q n}, k^{2 q m-(2 q-1) n}, A_{q}, B_{q}\right), q \in \mathbb{N} .
$$

As $m<n$, this means that after using the coxeter functor $C^{+}$a finite number of times, the modules generated will contain vector spaces of negative dimension, imaginary vector spaces so to say. The coxeter functor will only transform an indecomposable module into such "imaginary modules" if it is used on one of the projective modules, $P_{1}$ or $P_{2}$. So at the last point before the dimensions turn
negative, the module must have been reduced to one of the projective modules. Thus, we arrive at equation 2.9 if the projective module obtained is $P_{1}$ and equation 2.10 if the projective module obtained is $P_{2}$ :

$$
\begin{align*}
& \Phi^{q}\binom{m}{n}=\binom{0}{1}  \tag{2.9}\\
& (2 q+1) m-2 q n=0, \\
& 2 q m-(2 q-1) n=1 \\
& \Rightarrow \underline{n=m+1} \\
& \Phi^{q}\binom{m}{n}=\binom{1}{2}  \tag{2.10}\\
& (2 q+1) m-2 q n=1, \\
& 2 q m-(2 q-1) n=2 \\
& \Rightarrow \underline{n=m+1}
\end{align*}
$$

Proposition 2.8. Let $J=\left(k^{m}, k^{n}, A, B\right)$ be an indecomposable $\Lambda$-module. If $m>n$, then $m=n+1$

Proof: The proof is similar to the proof of Proposition 2.7, but instead of using the coxeter functor $C^{+}$, we may look at what happens when applying $C^{-}$ to an indecomposable $\Lambda$-module, $J$. Assume $J$ is not equal to $I_{1}$ or $I_{2}$, in which case it would belong to the class $N_{1}$ of. $C^{-}$, will yield negative dimensions after being applied some finite number of times, as the dimensions are given by $\Phi^{-q}$, and $m>n$. Thus, as applying $C^{-}$to an indecomposable $\Lambda$-module only yields imaginary modules when being applied to one of the injective $\Lambda$-modules, $I_{1}$ or $I_{2}$. We arrive at equations 2.11 and 2.12 by the same reasoning as for proposition 2.7 .

$$
\begin{align*}
& \Phi^{-q}\binom{m}{n}=\binom{1}{0}  \tag{2.11}\\
& -(2 q-1) m+2 q n=1, \\
& -2 q m+2(q+1) n=0 \\
& \Rightarrow \underline{m=n+1} \\
& \Phi^{-q}\binom{m}{n}=\binom{2}{1}  \tag{2.12}\\
& -(2 q-1) m+2 q n=2, \\
& -2 q m+(2 q+1) n=1 \\
& \Rightarrow \underline{m=n+1}
\end{align*}
$$

### 2.4.2 Indecomposables in the class $N_{3}$

By now, the modules in the classes $N_{1}$ and $N_{2}$ have been shown to be the only indecomposable $\Lambda$-modules where $n \neq m$ only by relying on the theory of the coxeter functor. However, to obtain the proof of existence and uniqueness of the last class of indecomposable modules, we need to apply the $\operatorname{Ext}_{\Lambda}^{1}(-, x)$ functor described in section 1.5 . We start by looking at the well-behaved structure of the 2 -Kronecker quiver.

Proposition 2.9. For any indecomposable $\Lambda$-module,

$$
J=\left(k^{m}, k^{n}, A, B\right), n<2 m, m \neq 0
$$

there exists a projective presentation

$$
0 \rightarrow p_{2}^{(m, n)} \rightarrow p_{1}^{(m, n)} \rightarrow J \rightarrow 0
$$

where $p_{2}^{(m, n)}$ and $p_{1}^{(m, n)}$ are given by:

$$
\begin{aligned}
p_{2}^{(m, n)} & =\left(P_{2}\right)^{2 m-n} \\
p_{1}^{(m, n)} & =\left(P_{1}\right)^{m}
\end{aligned}
$$

Proof: Let $J=\left(k^{m}, k^{n}, A, B\right)$ be an indecomposable left $\Lambda$-module. Now construct

$$
\begin{aligned}
& p_{1}^{m, n}=\left(k^{m}, k^{2 m}, A^{\prime}=\binom{i_{m}}{0}, B^{\prime}=\binom{0}{i_{m}}\right), \\
& p_{2}^{m, n}=\left(0, k^{2 m-n}, 0,0\right),
\end{aligned}
$$

where $A^{\prime}$ and $B^{\prime}$ are inclusions into the first $m$ copies of $k$ and the last $m$ copies of $k$ respectively. $p_{2}^{m, n}$ and $p_{1}^{m, n}$ are direct sums of a finite number of copies of projective modules $P_{2}$ and $P_{1}$ respectively, and thus, $p_{2}^{m, n}$ and $p_{1}^{m, n}$ are also projective, by proposition 1.5 . This gives the projective presentation

$$
0 \rightarrow p_{2}^{m, n} \rightarrow p_{1}^{m, n} \rightarrow M \rightarrow 0
$$

more precisely given by:

where $i$ is inclusion into $\operatorname{ker}(A B) \subseteq k^{2 m}$.
Proposition 2.10. Let $M=\left(k^{m}, k^{m}, A, B\right)$ be a left $\Lambda$-module. Then $M$ belongs to indecomposable class $N_{3}$ or it is decomposable by Theorem 2.6.

Proof: We may prove this by mathematical induction:
$m=1$ :
For the case where $m=1$, consider the case of the module $K$ defined by equation 2.13 .

$$
\begin{equation*}
K=K_{(x, y)}=(k, k, x, y) . \tag{2.13}
\end{equation*}
$$

In the case where $(x, y)=(0,0)$, the module would decompose as

$$
K_{(0,0)}=(k, k, 0,0)=(k, 0,0,0) \oplus(0, k, 0,0),
$$

which satisfies Theorem 2.6. Now, assume $(x, y) \neq(0,0)$, as the module would otherwise be decomposable. To create a projective resolution of this module, consider the projective modules $P_{1}$ and $P_{2}$, and the exact sequence:

$$
0 \rightarrow P_{2} \rightarrow P_{1} \rightarrow K_{(x, y)} \rightarrow 0
$$

given by:


Now, define $K^{\prime}=K_{\left(x^{\prime}, y^{\prime}\right)}$. To calculate the $\operatorname{Ext}_{\Lambda}^{1}\left(K, K^{\prime}\right)$, consider the exact sequence

$$
0 \rightarrow \mathrm{H}\left(K, K^{\prime}\right) \rightarrow \mathrm{H}\left(P_{1}, K^{\prime}\right) \rightarrow \mathrm{H}\left(P_{2}, K^{\prime}\right) \rightarrow \mathrm{E}\left(K, K^{\prime}\right) \rightarrow 0
$$

where

$$
\begin{array}{r}
\mathrm{H}(A, B)=\operatorname{Hom}_{\Lambda}(A, B), \\
\mathrm{E}(A, B)=\operatorname{Ext}_{\Lambda}^{1}(A, B) .
\end{array}
$$

The homomorphisms including the projective modules, $\mathrm{H}\left(P_{1}, K^{\prime}\right)$ and $\mathrm{H}\left(P_{2}, K^{\prime}\right)$, are completely determined by the first and second vector space respectively, hence, the length of their homomorphism group are both equal to 1 (they are of k-dimension 1).
Also $E\left(P_{1}, K^{\prime}\right)=E\left(P_{2}, K^{\prime}\right)=0$.
By Proposition 1.6 .

$$
\ell\left(\mathrm{E}\left(K, K^{\prime}\right)\right)=\ell\left(\mathrm{H}\left(K, K^{\prime}\right)\right)
$$

$\mathrm{H}\left(K, K^{\prime}\right)$ is the set of homomorphisms such that the following diagram commutes:

this is only possible for $(x, y)=\gamma\left(x^{\prime}, y^{\prime}\right), \gamma \in k$. If $K \simeq K^{\prime}, \mathrm{H}\left(K, K^{\prime}\right) \simeq k$, if not $\mathrm{H}\left(K, K^{\prime}\right)=0$, and so $\ell\left(\mathrm{E}\left(K, K^{\prime}\right)\right)=1$ or 0 respectively. Thus, the module may be written as $K_{(1, \lambda)} \simeq K_{(x, y)}$ by

$$
\lambda= \begin{cases}x^{-1} y & \text { if } x \neq 0 \\ \infty & \text { if } x=0\end{cases}
$$

Hence, for $m=1$, the module belongs to the indecomposable class $N_{3}$. $m=n$ :
Now assume that the statement holds for $m=n-1$. Let

$$
M_{n}=\left(k^{n}, k^{n}, A, B\right)
$$

be an indecomposable matrix. If both $A$ and $B$ are of full rank, this module is isomorphic to

$$
M_{n}^{*}=\left(k^{n}, k^{n}, I, A^{-1} B\right),
$$

and we may change the bases of $V_{1}$ and $V_{2}$ such that $A^{-1} B$ is on Jordan canonical form[1, Theorem 5.4, p. 423]. Thus the module belongs to $N_{3}$.
If $A$ is not of full rank, then $k \subseteq \operatorname{ker} A$, such that we have the exact sequence:

$$
0 \rightarrow S_{\infty}^{*} \rightarrow M_{n} \rightarrow L \rightarrow 0
$$

explicitly given by:

where $i$ is the inclusion, $p$ is the projection onto the quotient spaces of smaller dimension, and $\left.B\right|_{k} \simeq 1_{k}$.
By the induction hypothesis, the module

$$
L=\left(k^{n-1}, k^{n-1}, A^{\prime}, B^{\prime}\right)
$$

is either indecomposable or on the form of Theorem 2.6. Hence,

$$
\begin{aligned}
L & \simeq\left(k^{m_{1}}, k^{m_{1}+1},\binom{i_{m_{1}}}{0},\binom{0}{i_{m_{1}}}\right) \oplus \cdots \oplus\left(k^{m_{r}}, k^{m_{r}+1},\binom{i_{m_{r}}}{0},\binom{0}{i_{m_{r}}}\right) \\
& \oplus\left(k^{n_{1}+1}, k^{n_{1}},\left(i_{n_{1}} 0\right),\left(0 i_{n_{1}}\right)\right) \oplus \cdots \oplus\left(k^{n_{s}+1}, k^{n_{s}},\left(i_{n_{s}} 0\right),\left(0 i_{n_{s}}\right)\right) \\
& \oplus\left(k^{l_{1}}, k^{l_{1}}, i_{l_{1}}, J B_{l_{1}}^{\lambda_{1}}\right) \oplus \cdots \oplus\left(k^{l_{t}}, k^{l_{t}}, i_{l_{t}}, J B_{l_{t}}^{\lambda_{t}}\right), r, s, t \in \mathbb{N}, \\
& \lambda_{t} \in k \cup\{\infty\}
\end{aligned}
$$

Here

$$
\sum_{i=1}^{r} m_{i}=\sum_{j=1}^{s} n_{j}
$$

as the vector spaces have the same dimension.
As

$$
\mathrm{E}\left(\left(k^{m_{1}}, k^{m_{1}+1},\binom{i_{m_{1}}}{0},\binom{0}{i_{m_{1}}}\right),(k, k, 0,1)\right)=0,
$$

and

$$
\mathrm{E}\left(\left(k^{l_{i}}, k^{l_{i}}, I_{l_{i}}, J B_{l_{i}}^{\lambda_{i}}\right),(k, k, 0,1)\right)=0, \forall \lambda_{i} \neq \infty
$$

any such summands on the form of $N_{2}$, or $N_{3}$ with eigenvalues $\lambda_{i} \neq \infty$ would make the initial short sequence split, by Proposition 1.7, and hence, $M_{n} \simeq L \oplus(k, k, 0,1)$, which is a direct sum on the form of theorem 2.6

By this point it has been shown that the module $M_{n}$ is either on the form of Theorem 2.6 , or it is indecomposable, and

$$
L=\left(k^{s_{1}}, k^{s_{1}}, J B_{s_{1}}^{0}, i_{s_{1}}\right) \oplus \cdots \oplus\left(k^{s_{t}}, k^{s_{t}}, J B_{s_{t}}^{0}, i_{s_{t}}\right)
$$

Assume that the module $M_{n}$ is indecomposable. Thus, we may write the exact sequence explicitly by:


As each $i_{s_{u}}$ is of full rank for all $u \in\{1, \ldots, t\}$, and 1 is of full rank, this implies that $\operatorname{ker}\left(\oplus_{u=1}^{t} s_{i_{u}}\right)=0$, ker $1=0$, hence ker $B=0$. Thus, as $B$ is of full rank, we may apply a change of basis in $k^{n}$, such that

$$
M_{n} \simeq\left(k^{n}, k^{n}, B^{-1} A, I\right)
$$

where $B^{-1} A$ may be conjugated to be on the Jordan canonical form by a simultaneous change of basis of $V_{1}$ and $V_{2}$. This Jordan canonical form can not contain more than one Jordan block, as $M_{n}$ is assumed to be indecomposable, and hence, the Jordan canonical form of $B^{-1} A$ must be the single Jordan block with eigenvalue $\lambda=0$, as the matrix $B^{-1} A$ would otherwise be of full rank, and it is assumed that ker $A \neq 0$. Thus, the indecomposable matrix $M$ is on the form of $N_{3}$. A similar approach will yield the remaining modules of the class $N_{3}$ by assuming that $B$ is not of full rank.

## Chapter 3

## Systems of linear differential equations

In this chapter, we use the decomposition of representations of the 2-Kronecker quiver, $Q$ over an algebraically closed field, $k$, by theorem 2.6 to try to find a way to make it easier to solve systems of linear differential equations.

### 3.1 Systems of differential equations

A system of homogenous linear differential equations are systems on the form of equation 3.1 .

$$
\begin{equation*}
\frac{d}{d t} \mathbf{x}=\mathbf{A} \mathbf{x} \tag{3.1}
\end{equation*}
$$

For a system on this form, obtaining the solutions of the system is pretty straightforward, see for instance [4, p. 164-168]. However, the initial system may not be as well behaved. Consider the system in equation 3.2

$$
\begin{equation*}
\mathbf{A x}=\mathbf{B} \dot{\mathbf{x}} \tag{3.2}
\end{equation*}
$$

$\mathbf{A}$ and $\mathbf{B}$ are of the same dimension, $m \times n$, but otherwise, general matrices without any initial restrictions, thus, this system seems too general to be easily solved. Nevertheless, there are are some ways of manipulating A and $\mathbf{B}$ which will make us able to put the matrices on some very specific forms.

Proposition 3.1. Let $\boldsymbol{A}$ and $\boldsymbol{B}$ be $m \times n$-matrices. Let $\boldsymbol{P}$ be an invertible $m \times m$-matrix. Then solving the system

$$
\begin{equation*}
A x=B \dot{x} \tag{3.3}
\end{equation*}
$$

is equivalent to solving the system

$$
\begin{equation*}
P A x=P B \dot{x} \tag{3.4}
\end{equation*}
$$

Proof: If equation 3.3 has a solution, the same solution will still solve the system when applying the same invertible matrix on both sides of the equation. If equation 3.4 has a solution, as $\mathbf{P}$ is invertible, we may apply the same invertible matrix, $\mathbf{P}^{-1}$, to both sides of the equation while maintaining the same solution, hence:

$$
\mathbf{P}^{-1} \mathbf{P A x}=\mathbf{P}^{-1} \mathbf{P B} \dot{\mathbf{x}} \Rightarrow \mathbf{A x}=\mathbf{B} \dot{\mathbf{x}}
$$

and thus, equation 3.3 has a solution.
Proposition 3.2. Let $\boldsymbol{A}$ and $\boldsymbol{B}$ be $m \times n$-matrices. Let $\boldsymbol{Q}$ be an invertible $n \times n$-matrix. Then solving the system

$$
\begin{equation*}
\boldsymbol{A} \boldsymbol{x}=\boldsymbol{B} \dot{\boldsymbol{x}} \tag{3.5}
\end{equation*}
$$

is equivalent to solving the system

$$
\begin{equation*}
A Q y=B Q \dot{y} \tag{3.6}
\end{equation*}
$$

Proof: For any matrix M, we have that differentiating a vector, $\mathbf{x}$, and then applying the matrix is the same as applying the matrix and then differentiating, $\mathbf{M} \dot{\mathbf{x}}=(\mathbf{M} \mathbf{x})$. Let $\mathbf{Q}$ be an invertible $n \times n$ matrix, and let $\mathbf{A}$ and $\mathbf{B}$ be $m \times n$ matrices. The following systems are thus equivalent:

$$
\begin{align*}
\mathbf{A x} & =\mathbf{B} \dot{\mathbf{x}}  \tag{3.7}\\
\mathbf{A Q Q}^{-1} \mathbf{x} & =\mathbf{B Q Q}^{-1} \dot{\mathbf{x}} \\
\mathbf{A Q Q}^{-1} \mathbf{x} & =\mathbf{B Q}\left(\mathbf{Q}^{-1} \mathbf{x}\right) \\
\mathbf{A Q} \mathbf{y} & =\mathbf{B Q} \dot{\mathbf{y}} \tag{3.8}
\end{align*}
$$

The last system is obtained by the substitution $\mathbf{x}=\mathbf{Q y}$, and hence, we have solutions to the system of equation 3.6 if and only if we have solutions to the system of equation 3.5. The matrix describing the relationship between $\mathbf{x}$ and $\mathbf{y}$ is an invertible matrix, which is an isomorphism, thus the solutions for $\mathbf{x}$ are isomorphic to the solutions for $\mathbf{y}$.

Proposition 3.3. Let $\boldsymbol{A}$ and $\boldsymbol{B}$ be linear transformations from $k^{n}$ to $k^{m}$. Then the solutions to the system

$$
\boldsymbol{A x}=\boldsymbol{B} \dot{\boldsymbol{x}}
$$

is independent of the choice of basis for $k^{n}$ and $k^{m}$.

Proof: This follows immediately from proposition 3.1 and proposition 3.2, as a change of basis in $k^{l}$ corresponds to multiplication by an invertible $l \times l$-matrix.
Remark. Let A and $\mathbf{B}$ be the matrices in a system on the form of equation 3.2 As the solutions of a system is independent of the choice of basis, we may change our bases to obtain the matrices $\mathbf{A}^{\prime}$ and $\mathbf{B}^{\prime}$ on very specific forms, and solve the system given by these matrices instead. This is an interesting approach in theory, although actually finding the changes of bases required may be difficult in practice.

### 3.2 Matrix decomposition

Definition. Let A and Be be $m \times n$-matrices describing linear transformations from a vector space $k^{n}$ to a vector space $k^{m}$. The matrices are simultaneously decomposable if there exists a change of bases in $k^{n}$ and $k^{m}$ such that we may write $\mathbf{A} \simeq \mathbf{A}_{1} \oplus \mathbf{A}_{2}$ and $\mathbf{B} \simeq \mathbf{B}_{1} \oplus \mathbf{B}_{2}$, where $\mathbf{A}_{i}$ and $\mathbf{B}_{i}$ are both $m_{i} \times n_{i}$-matrices for $i \in\{1,2\}$.

Remark. A direct sum of matrices, $\mathbf{A}=\mathbf{A}_{1} \oplus \mathbf{A}_{2}$ is a block diagonal matrix such that:

$$
\mathbf{A}=\left(\begin{array}{c|c}
\mathbf{A}_{1} & 0 \\
\hline 0 & \mathbf{A}_{2}
\end{array}\right)
$$

Remark. Let $\mathbf{A}$ be a linear transformation between vector spaces $V_{1}$ and $V_{2}$. For the remainder of this chapter, we will use the terminology that a pair of matrices $(\boldsymbol{A}, \boldsymbol{B})$ contains a summand $\left(\boldsymbol{A}_{1}, \boldsymbol{B}_{1}\right)$ to describe that $\mathbf{A}$ and $\mathbf{B}$ are simultaneously decomposable in such a way that $\mathbf{A}_{1}$ is a summand of $\mathbf{A}$, and $\mathbf{B}_{1}$ is a summand of $\mathbf{B}$.

Let $\mathbf{A}$ and $\mathbf{B}$ be $m \times n$-matrices, that is, linear maps from a vector space $k^{n}$ to a vector space $k^{m}$. Now, the 4 -tuple $\left(k^{n}, k^{m}, A, B\right)$ corresponds to a representation of the quiver $Q$ over a field $k$, and hence, this representation is on the form given by theorem 2.6. As this is the case, we see that the pair of
matrices $(\mathbf{A}, \mathbf{B})$ contains a finite number of summands, each summand being an element in one of three classes $D_{1}, D_{2}$ or $D_{3}$ :

$$
\begin{aligned}
& D_{1}=\left\{\left(i_{n} 0\right),\left(0 i_{n}\right)\right\}, n \in \mathbb{N} \\
& D_{2}=\left\{\binom{i_{m}}{0},\binom{0}{i_{m}}\right\}, m \in \mathbb{N} \\
& D_{3}=\left\{i_{l}, J B_{l}^{\lambda}\right\}, l \in \mathbb{N}, \lambda \in k \cup\{\infty\},
\end{aligned}
$$

where $\lambda=\infty$ corresponds to the pair $J B_{l}^{0}, i_{l}$.
In other words, there exists a way to simultaneously decompose $\mathbf{A}$ and $\mathbf{B}$, such that instead of solving the system

$$
\mathbf{A x}=\mathbf{B} \dot{\mathbf{x}}
$$

we can choose to solve a finite number of systems on the forms:

$$
\begin{aligned}
\left(i_{n} 0\right) \mathbf{x}_{n} & =\left(0 i_{n}\right) \dot{\mathbf{x}}_{n}, n \in \mathbb{N}, \\
\binom{i_{m}}{0} \mathbf{x}_{m} & =\binom{0}{i_{m}} \dot{\mathbf{x}}_{m}, m \in \mathbb{N}, \\
i_{l} \mathbf{x}_{l} & =J B_{l}^{\lambda} \dot{\mathbf{x}}_{l}, l \in \mathbb{N} .
\end{aligned}
$$

Here, $\mathbf{x}_{i}$ denotes the vector of unknowns of dimension $i \times 1, \forall i \in \mathbb{N}$. We will consider what happens to the partial systems containing each of these classes of summands in order to get information about the general system.

### 3.3 Preliminary considerations

Let $(\mathbf{A}, \mathbf{B})$ be a pair of $m \times n$-matrices.

1. If $n>m$, the pair of matrices must contain direct summands in the class $D_{1}$.
2. If $m>n$, the pair of matrices must contain direct summands in the class $D_{2}$.

Thus, if we show that one of these summands yields systems that are unsolveable or have infinitely many solutions, the same will be the case for any pair of matrices containing such a summand.
For the case $m=n$, there are three different alternatives:
a) The pair of matrices may be written as a direct sum of an equal number of summands in the classes $D_{1}$ and $D_{2}$.
b) The pair of matrices may be written as a direct sum of summands in the class $D_{3}$ alone.
c) The pair of matrices may be written as a direct sum of an equal number of summands in the classes $D_{1}$ and $D_{2}$, and summands in the class $D_{3}$.

As either one of these may be the case, the fact that the matrices are square matrices does not provide specific information about the summands of the pair of matrices.

### 3.4 Direct summand in the class $D_{1}$

The direct summands in the first class, $D_{1}$, corresponding to modules on the form of class $N_{1}$, yields systems of infinitely many solutions, as information about one of the unknows are lost due to the matrices having rank less than the dimension of the codomain. We can see this by solving the system for a summand in the class $D_{1}$ explicitly:

$$
\begin{aligned}
& \mathbf{A x}=\mathbf{B} \dot{\mathbf{x}} \Rightarrow\left(\begin{array}{ll}
i_{n} & 0
\end{array}\right)\left(\begin{array}{c}
x_{1} \\
x_{2} \\
\vdots \\
x_{n+1}
\end{array}\right)=\left(\begin{array}{ll}
0 & i_{n}
\end{array}\right)\left(\begin{array}{c}
x_{1}^{\prime} \\
x_{2}^{\prime} \\
\vdots \\
x_{n+1}^{\prime}
\end{array}\right) \Rightarrow\left(\begin{array}{c}
x_{1} \\
x_{2} \\
\vdots \\
x_{n}
\end{array}\right)=\left(\begin{array}{c}
x_{2}^{\prime} \\
x_{3}^{\prime} \\
\vdots \\
x_{n+1}^{\prime}
\end{array}\right) \\
& \Rightarrow \text { System dependent upon unknown } x_{n+1}
\end{aligned}
$$

$\Rightarrow$ Infinitely many solutions to the system. No information about $x_{n+1}$.
Thus, if the initial pair of matrices contains direct summands corresponding to the class $N_{1}$, we can not have a uniquely determined solution to any initial value problem on this form.

### 3.5 Direct summand in the class $D_{2}$

Direct summands in the second class, $D_{2}$ yields only trivial solutions, as

$$
\mathbf{A} \mathbf{x}=\mathbf{B} \dot{\mathbf{x}} \Rightarrow\binom{i_{n}}{0}\left(\begin{array}{c}
x_{1} \\
x_{2} \\
\vdots \\
x_{n}
\end{array}\right)=\binom{0}{i_{n}}\left(\begin{array}{c}
x_{1}^{\prime} \\
x_{2}^{\prime} \\
\vdots \\
x_{n}^{\prime}
\end{array}\right)
$$

$$
\Rightarrow\left(\begin{array}{c}
x_{1} \\
x_{2} \\
\vdots \\
x_{n} \\
0
\end{array}\right)=\left(\begin{array}{c}
0 \\
x_{1}^{\prime} \\
\vdots \\
x_{n-1}^{\prime} \\
x_{n}^{\prime}
\end{array}\right) \Rightarrow \mathbf{x}=\overrightarrow{0}
$$

Hence, the summands in this class yield only trivial solutions to their part of the differential equation.

### 3.6 Direct summand in the class $D_{3}$

The third class of direct summands, summands in the class $D_{3}$, gives us completely determined systems for any eigenvalue, $\lambda \in k \backslash\{0\}$.

$$
\begin{aligned}
i_{n} \mathbf{x} & =\left(\begin{array}{ccccc}
\lambda & 1 & 0 & \cdots & 0 \\
0 & \lambda & 1 & \cdots & 0 \\
\vdots & \vdots & \ddots & \vdots \\
0 & 0 & \ddots & \vdots \\
0 & 0 & 0 & \cdots & \lambda
\end{array}\right) \dot{\mathbf{x}} \Rightarrow\left(\begin{array}{c}
x_{1} \\
\vdots \\
x_{n-1} \\
x_{n}
\end{array}\right)=\left(\begin{array}{c}
\lambda x_{1}^{\prime}+x_{2}^{\prime} \\
\vdots \\
\lambda x_{n-1}^{\prime}+x_{n}^{\prime} \\
\lambda x_{n}^{\prime}
\end{array}\right) \\
\mathbf{x} & =\left(\begin{array}{cc}
C_{1} t^{n-1} e^{\lambda t}+C_{2} t^{n-2} e^{\lambda t}+\cdots+C_{n} e^{\lambda t} \\
C_{2} t^{n-2} e^{\lambda t}+C_{3} t^{n-3} e^{\lambda t}+\cdots+C_{n} e^{\lambda t} \\
\vdots \\
& C_{n-1} t e^{\lambda t}+C_{n} e^{\lambda t} \\
C_{n} e^{\lambda t}
\end{array}\right)
\end{aligned}
$$

Here, the coefficients $C_{i}$ for $i \in\{1, \ldots, n\}$ corresponds to solutions of initial value problems.
For the eigenvalue $\lambda=0$, the system "shifts" the position of the derivatives, such as for summands in the class $D_{1}$, except for the last coordinate, which instead of an unknown becomes zero, and the solutions become the trivial solution.

$$
i_{n} \mathbf{x}=\left(\begin{array}{ccccc}
0 & 1 & 0 & \cdots & 0 \\
0 & 0 & 1 & \cdots & 0 \\
\vdots & \vdots & \ddots & \vdots \\
0 & 0 & 0 & \cdots & 1 \\
0 & 0 & 0 & \cdots & 0
\end{array}\right) \dot{\mathbf{x}} \Rightarrow\left(\begin{array}{c}
x_{1} \\
\vdots \\
x_{n-1} \\
x_{n}
\end{array}\right)=\left(\begin{array}{c}
x_{2}^{\prime} \\
\vdots \\
x_{n}^{\prime} \\
0
\end{array}\right) \Rightarrow \mathbf{x}=\overrightarrow{0} .
$$

For the eigenvalue $\lambda=\infty$, all the information available is that every value is dependent upon an unknown, $x_{1}$, where $x_{1}^{(n)}=0$.

$$
\left.\begin{array}{c}
\left(\begin{array}{cccc}
0 & 1 & \cdots & 0 \\
0 & 0 & 1 & \cdots \\
0 \\
\vdots & \vdots & \ddots & 0 \\
0 & 0 & \ddots & \vdots \\
0 & 0 & 0 & \cdots
\end{array}\right) \mathbf{1}
\end{array}\right) \mathbf{x}=i_{n} \dot{\mathbf{x}} \Rightarrow\left(\begin{array}{c}
x_{2} \\
\vdots \\
x_{n} \\
0
\end{array}\right)=\left(\begin{array}{c}
x_{1}^{\prime} \\
\vdots \\
x_{n-1}^{\prime} \\
x_{n}^{\prime}
\end{array}\right) . ~ \begin{gathered}
a_{n} \text { is a polynomial of degree at most } n-1 . \\
\quad \mathbf{x}=\left(\begin{array}{c}
a_{1} x+a_{2} \\
\vdots \\
a_{1} x^{n-1}+\cdots+a_{n-1} x+a_{n}
\end{array}\right)
\end{gathered}
$$

However, even as we see that the summands in the class $D_{3}$ of the system is completely solveable in theory, this approach assumes that we are able to find the decomposition of the initial pair of matrices such that we find all summands in the class $D_{3}$. This is the same as assuming that we are able to put an arbitrary matrix, M, on Jordan canonical form.

### 3.6.1 Rational canonical form

In order to obtain a Jordan canonical form of a matrix, M, we must be able to find it's eigenvalues, which may be very hard in principle. When dealing with pairs of matrices, $(\mathbf{A}, \mathbf{B})$, containing summands in the class $D_{3}$, we may use this information to construct another kind of simultaneous decomposition of $\mathbf{A}$ and B. As the pair $(\mathbf{A}, \mathbf{B})$ contains a summand in the class $D_{3}$, we know that there exists directs summands $\mathbf{A}^{\prime}$ in $\mathbf{A}$ and $\mathbf{B}^{\prime}$ in $\mathbf{B}$, such that $\mathbf{A}^{\prime}$ and $\mathbf{B}^{\prime}$ are two square matrices. We also know that at least one of $\mathbf{A}^{\prime}$ and $\mathbf{B}^{\prime}$ is of full rank.

## Assuming $\mathbf{B}^{\prime}$ is of full rank

Assume $\mathbf{B}^{\prime}$ is of full rank. Then solving the system

$$
\mathbf{A}^{\prime} \mathbf{x}=\mathbf{B}^{\prime} \dot{\mathbf{x}}
$$

is the same as solving the system

$$
\mathbf{A}^{\prime \prime} \mathbf{x}=\mathbf{B}^{\prime \prime} \dot{\mathbf{x}}
$$

where $\mathbf{A}^{\prime \prime}$ and $\mathbf{B}^{\prime \prime}$ are matrices such that $\mathbf{B}^{\prime \prime}$ is on the form of an identity matrix, and matrix $\mathbf{A}^{\prime \prime}$ is the rational canonical form of $\mathbf{A}^{\prime}{ }^{1}$ By this construction, we avoid the problem of obtaining a matrix on the form of a Jordan block. As the rational canonical form is obtained through conjugation by a pair of matrices $\mathbf{M}, \mathbf{M}^{-1}$, the change of basis required to put $\mathbf{A}^{\prime}$ in the rational canonical form will not distort the identity matrix, as the identity matrix is preserved under conjugation. Assume that the minimal polynomial $m(x)$ of $\mathbf{A}^{\prime}$, is equal to the characteristic polynomial $f(x)$ of $\mathbf{A}^{\prime}$. When this is the case, the rational canonical form is in the form of the companion matrix of $f(x)$. Thus, we get the system of linear differential equations as follows:

$$
\left(\begin{array}{ccccc}
0 & 0 & 0 & \cdots & 0 \\
1 & 0 & 0 & -a_{0} \\
0 & 1 & 0 & 0 & -a_{1} \\
\vdots & \vdots & \ddots & 0 & -a_{2} \\
\vdots & 0 & 0 & \vdots & \vdots \\
0 & 0 & \cdots & -a_{n-1}
\end{array}\right) \mathbf{x}=i_{n} \dot{\mathbf{x}} \Rightarrow\left(\begin{array}{c}
-a_{0} x_{n} \\
x_{1}-a_{1} x_{n} \\
\vdots \\
x_{n-1}-a_{n-1} x_{n}
\end{array}\right)=\left(\begin{array}{c}
x_{1}^{\prime} \\
x_{2}^{\prime} \\
\vdots \\
x_{n}^{\prime}
\end{array}\right)
$$

This gives the recursive relations:

$$
\begin{gathered}
x_{n-1}=a_{n-1} x_{n}+x_{n}^{\prime} \\
\Rightarrow x_{n-2}=a_{n-2} x_{n}+x_{n-1}^{\prime}=a_{n-2} x_{n}+a_{n-1} x_{n}^{\prime}+x_{n}^{\prime \prime} \\
\mathbf{x}=\left(\begin{array}{c}
x_{n}^{(n-1)}+\sum_{i=1}^{n-1} a_{n-i} x_{n}^{(n-1-i)} \\
\vdots \\
a_{n-1} x_{n}+x_{n}^{\prime} \\
x_{n}
\end{array}\right)
\end{gathered}
$$

Considering $x_{1}$ in this way, we get:

$$
x_{1}^{\prime}=-a_{0} x_{n}=\frac{d}{d t}\left(x_{n}^{(n-1)}+\sum_{i=1}^{n-1} a_{n-i} x_{n}^{(n-1-i)}\right)
$$

Which shows that if $a_{0}=0, x_{n}$ is constant, and this corresponds to the solutions of the system with summands in the class $D_{3}$ where $\lambda=\infty$. If $a_{0} \neq 0,0$ is not a root in the characteristic polynomial of $\mathbf{A}^{\prime}$, and thus, $\mathbf{A}^{\prime}$ is of full rank, as the eigenvalue of the Jordan block corresponding to $\mathbf{A}^{\prime}$ is non-zero. If this is the case, these solutions are included in the solutions obtained in the next part.

[^0]
## Assuming $\mathbf{A}^{\prime}$ is of full rank

Assume instead that $\mathbf{A}^{\prime}$ is of full rank. Now a change of basis in the vector spaces gives us a matrix $\mathbf{A}^{\prime \prime}$ which is an identity matrix, and a matrix $\mathbf{B}^{\prime \prime}$ which is the rational canonical form of $\mathbf{B}^{\prime}$. By assuming that the minimal polynomial of $\mathbf{B}^{\prime}$ is equal to the characteristic polynomial of $\mathbf{B}^{\prime}$, we get the system:

$$
\begin{aligned}
i_{n} \mathbf{x}=\left(\begin{array}{ccccc}
0 & 0 & 0 & \cdots & 0 \\
1 & 0 & -a_{0} \\
0 & 1 & 0 & \cdots & 0 \\
\hline & -a_{1} \\
\vdots & \vdots & \ddots & -a_{2} \\
0 & 0 & 0 & \cdots & \vdots \\
\vdots & -a_{n-1}
\end{array}\right) \dot{\mathbf{x}} \Rightarrow\left(\begin{array}{c}
x_{1} \\
x_{2} \\
\vdots \\
x_{n}
\end{array}\right)=\left(\begin{array}{c}
-a_{0} x_{n}^{\prime} \\
x_{1}^{\prime}-a_{1} x_{n}^{\prime} \\
\vdots \\
x_{n-1}^{\prime}-a_{n-1} x_{n}^{\prime}
\end{array}\right) \\
\Rightarrow x_{j}=-\sum_{i=0}^{j-1} a_{i} x_{n}^{(j-i)} \\
\Rightarrow x_{n}=-a_{n-1} x_{n}^{\prime}-\sum_{i=0}^{n-2} a_{i} x_{n}^{(n-i-2)}
\end{aligned}
$$

If $a_{0}=0$, the characteristic polynomial has a root $\lambda=0$, and as it is possible to put $\mathbf{B}^{\prime}$ on the form of a Jordan block, the characteristic polynomial of $\mathbf{B}^{\prime}$ has exactly one root. This implies that $a_{i}=0$ for all $i \in\{1, \ldots, n-1\}$, and thus, this yields only trivial solutions to the system, corresponding to summands in $D_{3}$ with eigenvalue $\lambda=0$. If $a_{0} \neq 0$, the recursive formula given, yields solutions of the form:

$$
\Rightarrow x_{j}=\sum_{i=0}^{j} C_{i} t^{(n-i)} e^{\gamma t}
$$

But as these solutions are isomorphic to the solutions given by solving the system with $\mathbf{B}^{\prime \prime}$ as a Jordan block, where $\lambda \in k \backslash\{0\}$, we deduce that finding the unknown, $\gamma$, corresponds to finding the eigenvalue of the Jordan block. Finding the solutions of the rational canonical form is equivalent to finding the eigenvalue of the Jordan block, and we have not managed to reduce the problem further.
Remark. Here, we assumed that the minimal polynomial was equal to the characteristic polynomial of the matrix in order to keep our computations simple. However, if the characteristic polynomial is not equal to the minimal polynomial, we obtain a block diagonal matrix which have blocks of the same form as the one we just computed. In this case, we would by the same reasoning deduce that solving the system corresponds to finding the eigenvalue of the Jordan block.

### 3.7 Final remarks

We have seen that the representations over the 2-Kronecker quiver is intimately related to the problem of decomposing a general system of linear differential equations. However, this approach is only feasible in principle, as obtaining a simultaneous decomposition of a pair of matrices is a highly non-trivial task in itself. Even if we were able to find the number of summands on the given form, and their dimensions, solving the system we are left with is equivalent to finding the eigenvalues of arbitrary quadratic matrices, which in turn is equivalent to finding all roots of the characteristic polynomial of an arbitrary matrix. So in conclusion, the connection discovered between solving systems of differential equations and the representations of the 2 -Kronecker quiver, though interesting and somewhat surprising, does not help us solve the systems of linear equations in general, as finding the roots of a general polynomial of degree higher than or equal to 5 is not possible, by the Abel-Ruffini theorem, see [3, p. 470].

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[^0]:    ${ }^{1}$ The definition of the rational canonical form of a matrix is omitted in this thesis, but the reader is referred to [1] for the definition used in this section.

