University of Utrecht Master Thesis Mathematical Sciences

# Dunkl operators and Fischer decompositions 

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#### Abstract

In this thesis we will study the theory of Dunkl operators and Dunkl harmonic polynomials and have a look at some of the applications. We will also establish the existence of a certain class of Fischer decompositions of graded vector spaces. The decomposition of $L^{2}\left(S, h^{2} d \omega\right)$ into Dunkl harmonics follows from a Fischer decomposition which belongs to this class. The similarities between Dunkl operators and partial derivatives can be expressed through a certain intertwining operator. The existence of this type of intertwining operator is explained from a general point of view.


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## Chapter 1

## Introduction

In 1917, E. Fischer (see [16]) introduced the following remarkable decompositions of the space $P=P\left(\mathbb{R}^{m}\right)$ of $\mathbb{R}$-valued polynomial functions on $\mathbb{R}^{m}$.
Let $j, n \in \mathbb{N}$ and let $f_{i} \in P, 1 \leq i \leq j$, be a set of linear independent homogeneous polynomials of degree $n$. Then we can decompose each polynomial $p(x)$ of degree $l>n$ as $p(x)=$ $q(x)+r(x)$, with $q$ contained in the ideal $\left(f_{1}, f_{2}, \ldots, f_{j}\right)=P f_{1}+\cdots+P f_{j}$ generated by $f_{1}, \ldots, f_{j}$, and $f_{i}\left(\partial_{x}\right) r(x)=0$, for $1 \leq i \leq j$. Here $f_{i}\left(\partial_{x}\right)$ is the element of the ring $\mathbb{R}\left[\partial_{1}, \ldots \partial_{m}\right]$ which is obtained from $f_{i}(x)$ by replacing each instance of $x_{j}$ with $\partial_{j}$.
A special case is the harmonic Fischer decomposition, which arises from $n=2, j=1, f_{1}=|x|^{2}$. By repeated use of this decomposition we can decompose $P$ as

$$
P=\bigoplus_{l \geq 0} \bigoplus_{i=0}^{\lfloor l / 2\rfloor}|x|^{2 i} H_{l-2 i}
$$

where $H_{l-2 i}$ is the space of harmonic polynomials which are homogeneous of degree $l-2 i$.
We will now introduce the notion of Dunkl operators. These were introduced in 1989 by Charles F. Dunkl [9] as a tool in his research on the orthogonal decomposition of $P$ with respect to an inner product defined in terms of a root system $R$ in $\mathbb{R}^{m}$. Let $R$ be such a root system, let $R_{+}$be a positive system and let $G$ be Weyl group. Let $k: R \rightarrow \mathbb{R}$ be a $G$-invariant function (a so-called weight function, we assume its values to be non-negative). For convenience we write $k_{\alpha}=k(\alpha)$. Let $h: \mathbb{R}^{m} \rightarrow \mathbb{R}$ be defined by

$$
h(x)=\prod_{\alpha \in R_{+}}|\langle\alpha, x\rangle|^{k_{\alpha}} .
$$

Let $S$ be the unit sphere in $\mathbb{R}^{m}$. Then the inner product on $P$ is defined by

$$
\langle p, q\rangle_{h}=\int_{S} p(x) q(x) h(x)^{2} d \omega
$$

with $d \omega$ the normalized rotation invariant measure on $S$. In other words, restriction to $S$ induces a linear injection from $P$ onto $\left.P\right|_{S} \subset L^{2}\left(S, h^{2} d \omega\right)$, and the inner product corresponds to the restriction of the square integrable inner product.
Associated with the fixed weight function $k$, the Dunkl operators are defined as follows, see 9].

$$
T_{i} f(x)=\partial_{i} f(x)+\sum_{\alpha \in R_{+}} k_{\alpha} \alpha_{i} \frac{f(x)-f\left(r_{\alpha}(x)\right)}{\langle x, \alpha\rangle} \text {, for } f \in C^{1}\left(\mathbb{R}^{m}\right)
$$

where $r_{\alpha}$ is the reflection in the hyperplane orthogonal to $\alpha$.
Note that these operators are the ordinary partial derivatives if $k=0$. Also note that $T_{i} f=\partial_{i} f$, if $f$ is $G$-invariant.
The Dunkl operators are homogeneous of degree -1 and they commute in $\operatorname{End}\left(C^{\infty}\left(\mathbb{R}^{m}\right)\right)$ for a fixed root system and weight function.
In view of those two properties one expects these operators to behave similar to partial derivatives. In fact, in [10] it was even shown that there exist an operator which intertwines the actions of the partial derivatives and Dunkl operators. We can use this intertwining operator to generalize many results from harmonic analysis to the setting of Dunkl operators. In particular we have a generalized Fourier transform and generalized Fischer decompositions. It will turn out that the generalized harmonic Fischer decomposition of $P$ naturally induces an orthogonal decomposition of $L^{2}\left(S, h^{2} d \omega\right)$.
There has also been some research on the application of Dunkl operators to physical systems (see [2], 6]) and recently the operators have been generalized to Clifford spaces (see [3, [4]). In Chapters 2 to 3, we will look at the harmonic Fischer decomposition of $P$ and we will give the explicit decomposition by use of harmonic analysis and some representation theory. In Chapter 4, we shall show the existence of Fischer decompositions in arbitrary graded vector spaces. In Chapter 5, we will show the existence of a certain class of Fischer decompositions of $P$. In Chapters 6 to 9 , we will review many of the results from Dunkl's papers including the construction of the above mentioned intertwining operator and the construction of the so called Dunkl transform. In Chapters 10 to 12 we will give some applications of this transform to certain types of differential-difference equations. Finally in Chapter 13, we will look at the existence of intertwining operators in graded vectors spaces. The results from this chapter will imply uniqueness of the intertwining operator between the Dunkl operators and the partial derivatives.

## Chapter 2

## The harmonic Fischer decomposition

In this chapter we are going to decompose the space of $\mathbb{R}$-valued polynomial functions on $\mathbb{R}^{m}$ as a direct sum of the vector spaces $|x|^{2 i} H_{j}, i, j \in \mathbb{N}, x \in \mathbb{R}^{m}$, where $H_{j}$ is the space of homogeneous harmonic polynomials of degree $j$ on $\mathbb{R}^{m}$. We shall do this by using some representation theory.

Denote by $P$ the space of $\mathbb{R}$-valued polynomial functions on $\mathbb{R}^{m}$. Let $P_{n}$ be the space of homogeneous polynomials of degree $n$ on $\mathbb{R}^{m}$. We have the decomposition

$$
P=\bigoplus_{n \in \mathbb{N}} P_{n},
$$

as direct sum of vector spaces.
Denote by $\Delta: C^{\infty}\left(\mathbb{R}^{m}\right) \rightarrow C^{\infty}\left(\mathbb{R}^{m}\right)$ the Laplacian given by

$$
\Delta=\sum_{i=1}^{m} \partial_{i}^{2}
$$

and denote by $E: C^{\infty}\left(\mathbb{R}^{m}\right) \rightarrow C^{\infty}\left(\mathbb{R}^{m}\right)$ the Euler operator given by

$$
E=\sum_{i=1}^{m} x_{i} \frac{\partial}{\partial x_{i}} .
$$

We will also use the multiplication by $|x|^{2}$, which maps $C^{\infty}\left(\mathbb{R}^{m}\right)$ into $C^{\infty}\left(\mathbb{R}^{m}\right)$.
Since $\partial_{i} P \subset P$ and $x_{i} P \subset P$, we can restrict $\Delta, E$ and $|x|^{2}$ to linear operators on $P$. Denote by $H_{n}=P_{n} \cap \operatorname{ker}(\Delta)$ the subspaces of harmonic polynomials. Also note that the spaces $P_{n}$ are the eigenspaces of $E$ and for $p \in P_{n}, E(p)=n p$.
The main theorem of this chapter is
Theorem 2.1. The space $P_{n}$ admits the decomposition

$$
P_{n}=H_{n} \oplus|x|^{2} H_{n-2} \oplus|x|^{4} H_{n-4} \oplus \ldots
$$

The decomposition in Theorem 2.1 is the harmonic Fischer decomposition and it is an example of the Fischer decompositions described by E. Fischer in [16]. We will prove the existence of this decomposition in a different way.
Before we can prove Theorem 2.1, we need some additional lemmas and some representation theory.

Lemma 2.2. [4, p. 2] The linear span of the operators $|x|^{2}, \Delta$ and $E+m / 2$ in $\operatorname{End}\left(C^{\infty}\left(\mathbb{R}^{m}\right)\right)$ equipped with the commutator bracket is a Lie algebra isomorphic to $\mathfrak{s l}_{2}$.

Proof. We need to check the commutation relations of these 3 operators. For $\Delta$ and $|x|^{2}$ we find

$$
\begin{aligned}
{\left[\Delta,|x|^{2}\right] f } & =\Delta|x|^{2} f-|x|^{2} \Delta f \\
& =2 \sum_{i=1}^{m} \partial_{i}\left(|x|^{2}\right) \partial_{i}(f)+\sum_{i=1}^{m} \partial_{i}^{2}\left(|x|^{2}\right) f \\
& =\sum_{i=1}^{m} 4 E(f)+2 m f \\
& =4(E+m / 2) f
\end{aligned}
$$

For $E+m / 2$ and $|x|^{2}$ we find

$$
\begin{aligned}
{\left[E+m / 2,|x|^{2}\right] f } & =\left(E|x|^{2}-|x|^{2} E\right) f \\
& =\sum_{i=1}^{m}\left(|x|^{2} x_{i} \partial_{i}-|x|^{2} x_{i} \partial_{i}+\partial_{i}\left(|x|^{2}\right) x_{i}\right) f \\
& =\sum_{i=1}^{m} 2 x_{i}^{2} f \\
& =2|x|^{2} f .
\end{aligned}
$$

For $E+m / 2$ and $\Delta$ we find

$$
\begin{aligned}
{[E+m / 2, \Delta,] f } & =(E \Delta-\Delta E) f \\
& =\sum_{i=1}^{m}\left(x_{i} \partial_{i}^{3}-x_{i} \partial_{i}^{3}-2 \partial_{i}\left(x_{i}\right) \partial_{i}^{2}\right) f \\
& =-2 \sum_{i=1}^{m} \partial_{i}^{2} f \\
& =-2 \Delta f
\end{aligned}
$$

so $|x|^{2}, \Delta$ and $E+m / 2$ span a Lie algebra isomorphic to $\mathfrak{s l}_{2}$.
Note that the rescaled operators $1 / 2|x|^{2}, 1 / 2 \Delta$ and $E+m / 2$ form a standard $\mathfrak{s l}_{2}$-triple. See also [4, p. 2].

Lemma 2.3. Let $k \neq 0 \in \mathbb{N}$. The operator $\Delta|x|^{2}$ acts as a nonzero scalar on each of the subspaces $|x|^{2 k} H_{n}$ of $P$. In particular, its action is invertible on $|x|^{2 k} H_{n}$.

Proof. First choose $h_{n} \in H_{n}$. Using the commutator relations, we see that

$$
\Delta|x|^{2} h_{n}=|x|^{2} \Delta h_{n}+4(E+m / 2) h_{n}=4(n+m / 2) h_{n} .
$$

We can compute $\Delta|x|^{2 k} h_{n}$, by using the Leibniz rule

$$
\left[A^{k}, E\right]=\sum_{j=1}^{k-1} A^{j-1}[A, E] A^{k-j}
$$

with $A=|x|^{2}$. This gives

$$
\begin{align*}
\Delta|x|^{2 k} h_{n} & =|x|^{2 k} \Delta h_{n}+\sum_{i=0}^{k-1}|x|^{2(k-1-i)}\left[|x|^{2}, E\right]|x|^{2 i} h_{n} \\
& =0+\sum_{i=0}^{k-1}|x|^{2(k-1-i)} 4(E+m / 2)|x|^{2 i} h_{n} \\
& =\sum_{i=0}^{k-1}|x|^{2(k-i-1)} 4(2 i+n+m / 2)|x|^{2} h_{n} \\
& =\sum_{i=0}^{k-1} 4(2 i+n+m / 2)|x|^{2(k-1)} h_{n} \\
& =4 k(n+m / 2+k-1)|x|^{2(k-1)} h_{n} \\
& :=c_{n k}|x|^{2(k-1)} h_{n} \tag{2.1}
\end{align*}
$$

where we have used that $\Delta h_{n}=0$ to get the second equality and we have used that $|x|{ }^{2 i} h_{n} \in$ $P_{n+2 i}$ to get the third equality.

The constants $c_{n k}$ are nonzero and they also depend on $m$. This dependence is omitted from the notation, because $m$ is fixed throughout the paper.

Corollary 2.4. For $h \in H_{n}$ we find

$$
\begin{aligned}
\Delta^{i}|x|^{2 k} h_{n} & =\left(\prod_{j=k-i+1}^{k} 4 j(n+m / 2+j-1)\right)|x|^{2 k-2 i} h_{n} \\
& =4^{i}(k-i+1)_{i}(n+m / 2+k-i)_{i}|x|^{2 k-2 i} h_{n} \\
& =4^{i}(-k)_{i}(-n-k-m / 2+1)_{i}|x|^{2 k-2 i} h_{n}
\end{aligned}
$$

Here we have used the notation $(i)_{j}=i \cdot(i+1) \cdots \cdots(i+j-1)$.
Proof. These constants are found by applying Lemma 2.3 repeatedly.
So the eigenspace decomposition of $\Delta|x|^{2}$ looks a lot like the harmonic Fischer decomposition, but we still need to show that each polynomial can be written as a sum of terms of the form $|x|^{2 k} h_{n}$.

## Proof of Theorem 2.1

We will use induction on $n$. Since all polynomials of degree 0 and 1 are harmonic, the decomposition is trivial for $n=0$ or $n=1$.
Let $n \geq 2$ and assume that the decomposition of $P_{k}$ holds for all $k \leq n-2$. We will show that $P_{n}$ can also be decomposed.
Take $p \in P_{n}$, then $\Delta p=q \in P_{n-2}$. We can use the harmonic Fischer decomposition of $P_{n-2}$, to get $q=q_{1}+|x|^{2} q_{2}+|x|^{4} q_{3}+\ldots$, with $q_{i} \in H_{n-2 i}$. Using Lemma 2.3. we find

$$
\Delta|x|^{2} q=\sum_{i} c_{n-2 i, i}|x|^{2 i-2} q_{i} .
$$

Define $q^{\prime}$ by

$$
q^{\prime}=\sum_{i}\left(c_{n-2 i, i}\right)^{-1}|x|^{2 i-2} q_{i},
$$

then $\Delta|x|^{2} q^{\prime}=q$.
This means that the polynomial $p-|x|^{2} q^{\prime}$ is harmonic and that the decomposition of $p$ is given by

$$
p=\left(p-|x|^{2} q^{\prime}\right)+|x|^{2} q^{\prime}=\left(p-|x|^{2} q^{\prime}\right)+\sum_{i} 1 / c_{n-2 i, i}|x|^{2 i} q_{i},
$$

which shows that

$$
P_{n}=H_{n}+|x|^{2} H_{n-2}+|x|^{4} H_{n-4}+\ldots .
$$

Next we need to prove that the decomposition is a direct sum of vector spaces. For this we need to show uniqueness of the coefficients $q_{i}$.
Again we will use induction on $n$. Note that the coefficients are unique for $n=0$ and $n=1$. Assume that the sum is direct on $P_{n-2}$. Next choose $p \in P_{n}$ arbitrary and assume

$$
p=a_{0}+\sum|x|^{2 i} a_{i}=b_{0}+\sum|x|^{2 i} b_{i},
$$

with $a_{i}$ and $b_{i}$ harmonic. Applying $\Delta$ to these equations gives

$$
\Delta p=\sum c_{n-2 i, i}|x|^{2 i-2} a_{i}=b_{0}+\sum c_{n-2 i, i}|x|^{2 i-2} b_{i} .
$$

Because the decomposition for $\Delta p$ is unique, we have that $a_{i}=b_{i}$ for $i>0$, so $a_{0}=b_{0}$ and the decomposition of $P_{n}$ is unique.

The following corollary is a special case of [8, p. 39].
Corollary 2.5. Let $S$ be unit sphere $\left\{x \in \mathbb{R}^{m}:|x|=1\right\}$ and let $B$ be the open unit ball $\left\{x \in \mathbb{R}^{m}:|x|<1\right\}$. Restriction of the harmonic Fischer decomposition defined in Theorem 2.1 to $S$ leads to the decomposition

$$
L^{2}\left(S, h^{2} d \omega\right)=\widehat{\bigoplus}_{n \in \mathbb{N}}^{\perp} H_{n} \mid S
$$

Proof. Let $n, m \in \mathbb{N}, n \neq m$. Let $p \in H_{n}, q \in H_{m}$. Let $\eta$ be the outward normal vector on $S$. We have that

$$
\begin{aligned}
0 & =\int_{\bar{B}}(\Delta p) q-p \Delta(q) d x \\
& =\int_{S}\left(\frac{d p}{d \eta} q-p \frac{d q}{d \eta}\right) d \omega \\
& =\int_{S} p q(\operatorname{deg}(p)-\operatorname{deg}(q)) d \omega \\
& =(n-m) \int_{S} p q d \omega
\end{aligned}
$$

where we have used Green's theorem to get the second equality. Since $n \neq m$, it follows that $\int_{S} p q d \omega=0$, so $\left.\left.H_{n}\right|_{S} \perp H_{m}\right|_{S}$ in $L^{2}(S, d \omega)$.
Since $|x|^{2 n}=1$ on the unit sphere, we have that

$$
\left.P\right|_{S}=\left.\sum_{n=0}^{\infty} \sum_{i=0}^{\lfloor n / 2\rfloor}\left(|x|^{2 i} H_{n-2 i}\right)\right|_{S}=\left.\left.P\left(|x|^{2}\right)\right|_{S} \bigotimes \sum_{n=0}^{\infty} H_{n}\right|_{S}=\left.\sum_{n=0}^{\infty} H_{n}\right|_{S},
$$

where we have denoted the space of all polynomials in $|x|^{2}$ by $P\left(|x|^{2}\right)$. Since $\left.\left.H_{n}\right|_{S} \perp H_{m}\right|_{S}$ for $n \neq m$, the sum $\left.\sum_{n=0}^{\infty} H_{n}\right|_{S}$ is orthogonal. In particular, it is direct.
By Stone-Weierstrass the space $\left.P\right|_{S}$ is dense $C(S)$, so it is also dense in $L^{2}(S, d \omega)$. This gives the decomposition

$$
L^{2}(S, d \omega)=\widehat{\bigoplus}_{n \in \mathbb{N}}^{\perp} H_{n} \mid S
$$

## Chapter 3

## Construction of a basis for $H_{n}$

In this chapter we will use the harmonic Fischer decomposition of Theorem 2.1 to construct a basis for each of the spaces $H_{n}(n \in \mathbb{N})$. This Fischer decomposition gives rise to projections $\pi_{n}: P_{n} \rightarrow H_{n}$, which project homogeneous polynomials to their harmonic parts in a natural way. Since $P_{n}$ has a much larger dimension then $H_{n}$, the main problem is to find a set of functions $f_{i}$, such that the functions $\pi_{n}\left(f_{i}\right)$ form a basis of $H_{n}$.
To solve this problem, we first need to determine the dimension of the space $H_{n}$. Next we will introduce spherical coordinates in more than 3 dimensions and finally we define the polynomials $\phi_{n k l}$ in these spherical coordinates, such that the functions $\pi_{n}\left(\phi_{n k l}\right)$ form a basis of $H_{n}$.

Lemma 3.1. Denote the dimension of $P_{n}\left(\mathbb{R}^{m}\right)$ by $p(m, n)$, then $\operatorname{dim}\left(H_{n}\right)$ is given by

$$
\operatorname{dim}\left(H_{n}\left(\mathbb{R}^{m}\right)\right)=p(m-1, n)+p(m-1, n-1)
$$

Proof. The dimension of $P_{n}\left(\mathbb{R}^{m}\right)$ is equal to $(n+m-1)$ choose $n$. To prove this we need to solve a counting problem.
Each partition $q_{l}$ of a sequence of $n+m$ elements into $m$ parts is given by an $m$-dimensional vector $l$. The number of possible partitions is given by $(n+m-1)$ choose $m-1$, since we need to choose $m-1$ positions where we split the sequence, out of the $m+n-1$ possible positions. There is a one-to-one correspondence between the monomials $x^{k}=x_{1}^{k_{1}} \ldots x_{m}^{k_{m}}$ of degree $|k|=n$ and the partitions $q_{l}$, given by $k_{i}=l_{i}=1,1 \leq i \leq m$, so $p(m, n)$ is equal to $n+m-1$ choose $n$. Here we have used the multi-index notation $x^{k}=x_{1}^{k_{1}} \ldots x_{m}^{k_{m}}$.
From the definition of $n$ choose $k$ it follows that $p(m, n)=p(m, n-1)+p(m-1, n)$ and from the harmonic Fischer decomposition it follows that $H_{n} \simeq P_{n} /\left(|x|^{2} P_{n-2}\right)$, so

$$
\begin{aligned}
\operatorname{dim}\left(H_{n}\right) & =p(m, n)-p(m, n-2) \\
& =p(m-1, n)+p(m, n-1)-p(m, n-2) \\
& =p(m-1, n)+p(m-1, n-1)+p(m, n-2)-p(m, n-2) \\
& =p(m-1, n)+p(m-1, n-1) .
\end{aligned}
$$

Definition 3.2. For $m \geq 2$, let $U$ be the open subset $(0, \infty) \times(0, \pi)^{m-2} \times(0,2 \pi)$ of $\mathbb{R}^{m}$ and denote the elements of $U$ by $\left(r, \theta_{1}, \ldots, \theta_{m-1}\right)$.

Define the function $\zeta: \bar{U} \rightarrow \mathbb{R}^{m}$ by

$$
\begin{align*}
\zeta_{i}\left(r, \theta_{1}, \ldots, \theta_{m-1}\right) & =r \prod_{k=1}^{i-1} \sin \left(\theta_{k}\right) \cos \left(\theta_{i}\right) \text { for } 1 \leq i<m  \tag{3.1}\\
\zeta_{m}\left(r, \theta_{1}, \ldots, \theta_{m-1}\right) & =r \prod_{k=1}^{m-1} \sin \left(\theta_{k}\right) \tag{3.2}
\end{align*}
$$

The numbers $\left(r, \theta_{1}, \ldots, \theta_{m-1}\right)$ are called the spherical coordinates of the point $\zeta\left(r, \theta_{1}, \ldots, \theta_{m-1}\right) \in$ $\mathbb{R}^{m}$.

Note that for $m=2$, Definition 3.2 gives the usual polar coordinates on $\mathbb{R}^{2}$ and for $m=3$ Definition 3.2 gives the usual spherical coordinates on $\mathbb{R}^{3}$.
Theorem 3.3. The map $\zeta: \bar{U} \rightarrow \mathbb{R}^{m}$ is surjective. Let $V$ be the open set $\mathbb{R}^{m} \backslash\left\{x \in \mathbb{R}^{m} \mid x_{m}=0, x_{m-1} \geq 0\right\}$. The map $\left.\zeta\right|_{U}: U \rightarrow V$ is a $C^{\infty}$ diffeomorphism.
Proof. For $x \in \mathbb{R}^{m}, x \neq 0$, write

$$
x=\sum_{i=1}^{m} x_{i} e_{i}
$$

with $e_{i}$ the $i^{\text {th }}$ standard basis vector of $\mathbb{R}^{m}$. Define

$$
r_{k}=\sqrt{\sum_{i=k}^{m} x_{i}^{2}}, \text { and } y_{k}=\sum_{i=k+1}^{m} x_{i} e_{i},
$$

so $y_{k}$ is the projection of $x$ onto the last $m-k$ coordinates and $r_{k}$ is the norm of $y_{k}$. Also note that $y_{k-1}=y_{k}+x_{k} e_{k}$, so $r_{k}-1 \geq r_{k}$. Also note that $r_{1}=|x|=r$ and $r_{m}=\left|x_{m}\right|$.
For $1 \leq k \leq m-2$, there is a unique $\theta_{k} \in[0, \pi]$, such that $x_{k}=r_{k} \cos \left(\theta_{k}\right)$ and $r_{k+1}=r_{k} \sin \left(\theta_{k}\right)$. This can be seen by looking at the $\left(e_{k}, y_{k}\right)$-plane and the triangle $\left(x_{k} e_{k}, y_{k-1}, 0\right)$ in this plane. There also is a unique $\theta_{m-1} \in[0,2 \pi)$, such that $x_{m-1}=r_{m-1} \cos \left(\theta_{m-1}\right)$ and $x_{m}=r_{m-1} \sin \left(\theta_{m-1}\right)$. This shows that $\zeta: \bar{U} \rightarrow \mathbb{R}^{m}$ is surjective.
Next assume that $\zeta\left(r, \theta_{1}, \ldots, \theta_{m-1}\right)=\zeta\left(r^{\prime}, \theta_{1}^{\prime}, \ldots, \theta_{m-1}^{\prime}\right)$, for two points in $\bar{U}$. Then either $r=r^{\prime}=0$ and $\theta_{i} \neq \theta_{i}^{\prime}$ for some $i$, or $\theta_{i}=\theta_{i}^{\prime}$, for $i \leq k<m, \theta_{k}=\theta_{k}^{\prime}=0$ or $\pi$ and $\theta_{i} \neq \theta_{i}$ for some $i>k$. However all those points are elements of $\bar{U} \backslash U$ so $\left.\zeta\right|_{U}$ is injective.
To show that $\zeta: U \rightarrow V$ is surjective, we need to compute its image. First look at the image of $\bar{U} \backslash U$ under $\zeta$. If $u \in \bar{U} \backslash U$, either $r=0$, or one the angles $\theta$ is 0 or $\pi$. In those cases $x_{m}=0$ and $x_{m-1} \geq 0$ by the positive of the sine on $(0, \pi)$. So $\zeta(\bar{U} \backslash U)=\left\{x \in \mathbb{R}^{m} \mid x_{m}=0, x_{m-1} \geq 0\right\}$. Suppose that for some $u \in U, \zeta(u) \in\left\{x \in \mathbb{R}^{m} \mid x_{m}=0, x_{m-1} \geq 0\right\}$. Then $\theta_{m-1}=\pi$ by the definition of $\zeta_{m}$. However this means that $\zeta_{m-1}<0$, which leads to a contradiction.
This shows that $\zeta: U \rightarrow V$ is a bijection.
Next we need to prove that the determinant of the total derivative $D \zeta$ is nonzero on $U$. Here we can even prove that $\operatorname{det}(D \zeta)=r^{m-1} \sin \left(\theta_{1}\right)^{m-2} \sin \left(\theta_{2}\right)^{m-3} \ldots \sin \left(\theta_{m-3}\right)^{2} \sin \left(\theta_{m-2}\right)^{1}$ by induction over $m$ and a direct computation.
We shall denote the total derivative $D \zeta$ by $J^{m}$ to show the $m$-dependence explicitly. Note that the superscript $m$ is an index and not a power. $J^{m}$ is a $m \times m$-matrix.
For $m=1$, we have $J^{1}=1$, so $\operatorname{det}\left(J^{1}\right)=1$
For $m=2$, we have

$$
J^{2}=\left(\begin{array}{ll}
\cos \left(\theta_{1}\right) & -r \sin \left(\theta_{1}\right) \\
\sin \left(\theta_{1}\right) & r \cos \left(\theta_{1}\right)
\end{array}\right),
$$

so $\operatorname{det}\left(J^{2}\right)=r$.
Now suppose that $\operatorname{det}\left(J^{M}\right)=r^{M-1} \sin \left(\theta_{1}\right)^{M-2} \sin \left(\theta_{2}\right)^{M-3} \ldots \sin \left(\theta_{M-3}\right)^{2} \sin \left(\theta_{M-2}\right)^{1}$ for some $M \in \mathbb{N}$. We need to show that $\operatorname{det}\left(J^{M+1}\right)=r^{M} \sin \left(\theta_{1}\right)^{M-1} \sin \left(\theta_{2}\right)^{M-2} \ldots \sin \left(\theta_{M-2}\right)^{2} \sin \left(\theta_{M-1}\right)^{1}$. To do this we write

$$
J^{M+1}=\left(\begin{array}{llll}
J_{11}^{M} & \ldots & J_{1 m}^{M} & 0 \\
\vdots & & & \vdots \\
J_{M-1,1}^{M} & \ldots & J_{M-1, m}^{M} & 0 \\
J_{M, 1}^{M} \cos \left(\theta_{M}\right) & \ldots & J_{M, m}^{M} \cos \left(\theta_{M}\right) & -r \sin \left(\theta_{M}\right) \prod_{i=1}^{M-1} \cos \left(\theta_{M}\right) \\
J_{M, 1}^{M} \sin \left(\theta_{M}\right) & \ldots & J_{M, m}^{M} \sin \left(\theta_{M}\right) & r \cos \left(\theta_{M}\right) \prod_{i=1}^{M-1} \cos \left(\theta_{M}\right)
\end{array}\right) .
$$

To compute the determinant of $J^{M+1}$, we expand the matrix along the $(M+1)^{t h}$ column, which gives

$$
\begin{aligned}
\operatorname{det}\left(J^{m+1}\right) & =\operatorname{det}\left(J^{m}\right)\left[\cos \left(\theta_{M}\right) \cdot r \cos \left(\theta_{M}\right) \prod_{i=1}^{M-1} \cos \left(\theta_{M}\right)-\sin \left(\theta_{M}\right) \cdot-r \sin \left(\theta_{M}\right) \prod_{i=1}^{M-1} \cos \left(\theta_{M}\right)\right] \\
& =r \operatorname{det}\left(J^{m}\right) \prod_{i=1}^{M-1} \cos \left(\theta_{M}\right) \\
& =r^{M} \sin \left(\theta_{1}\right)^{M-1} \sin \left(\theta_{2}\right)^{M-2} \ldots \sin \left(\theta_{M-2}\right)^{2} \sin \left(\theta_{M-1}\right)^{1},
\end{aligned}
$$

where we have used that $\sin ^{2}\left(\theta_{M}\right)+\cos ^{2}\left(\theta_{M}\right)=1$ to get the second equation. So by the induction hypothesis $\operatorname{det}(D \zeta)=r^{m-1} \sin \left(\theta_{1}\right)^{m-2} \sin \left(\theta_{2}\right)^{m-3} \ldots \sin \left(\theta_{m-3}\right)^{2} \sin \left(\theta_{m-2}\right)^{1}$, which is nonzero on all of $U$.
So $\zeta: U \rightarrow V$ is a $C^{\infty}$ diffeomorphism.
Lemma 3.4. Let $p(x)=x^{k}$ be a monomial of degree $|k|=n$ in Cartesian coordinates on $\mathbb{R}^{m}$, where we have used the multi-index notation $x^{k}=x_{1}^{k_{1}} \ldots x_{m}^{k_{m}}$, for $k \in \mathbb{N}^{m}$. We can rewrite $p(x)$ in spherical coordinates as

$$
\begin{align*}
& p \circ \zeta\left(r, \theta_{1}, \ldots, \theta_{m-1}\right)=\zeta\left(r, \theta_{1}, \ldots, \theta_{m-1}\right)^{k}= \\
& \quad r^{n} \prod_{i=1}^{m-2}\left(\sin \left(\theta_{i}\right)^{n-k_{1}-\cdots-k_{i}} \cos \left(\theta_{i}\right)^{k_{i}}\right) \sin \left(\theta_{m-1}\right)^{k_{m}} \cos \left(\theta_{m-1}\right)^{k_{m-1}} \in C^{\infty}(U) . \tag{3.3}
\end{align*}
$$

The function $p \circ \zeta$ can be extended to $C^{\infty}$-function on $\bar{U}$.
Proof. Formula (3.3) is proven by a direct computation using equations (3.1) and (3.2). We can extend $\zeta$ to a function from $\bar{U}$ to $\mathbb{R}^{m}$ in a natural way. For $x \in \mathbb{R}^{m}$ the preimage $\zeta^{-1}(x)$ is the set of points $u \in \bar{U}$, with $\zeta(u)=x$. By a direct computation we see that $p \circ \zeta(u)=p \circ \zeta\left(u^{\prime}\right)$, if both $u, u^{\prime} \in \zeta^{-1}(\zeta(u))$. Because the sine and cosine are smooth functions the extension of $p \circ \zeta$ to $\bar{U}$ is also a smooth function.

Definition 3.5. Let $n \in \mathbb{N}$ and let $I$ be the set $\left\{(k, l) \in \mathbb{Z}_{+}^{m-2} \times \mathbb{Z}\left|\sum k_{i}+|l|=n\right\}\right.$. Define the functions $\phi_{n, k, l}$ by

$$
\phi_{n, k, l}= \begin{cases}r^{n} \prod_{i=1}^{m-1}\left[\sin \left(\theta_{i}\right)^{n-k_{1} \cdots-k_{i-1}} \cos \left(\theta_{i}\right)^{k_{i}}\right] \cos \left(l \theta_{m-1}\right) ; & l \geq 0, \\ r^{n} \prod_{i=1}^{m-1}\left[\sin \left(\theta_{i}\right)^{n-k_{1} \cdots-k_{i-1}} \cos \left(\theta_{i}\right)^{k_{i}}\right] \sin \left(-l \theta_{m-1}\right) ; & l<0 .\end{cases}
$$

Define the linear spaces $\Phi_{n}$ by $\Phi_{n}=\operatorname{span}\left\{\phi_{n, k, l} \mid(k, l) \in I\right\}$.

Lemma 3.6. $\operatorname{dim}\left(\Phi_{n}\right)=\operatorname{dim}\left(H_{n}\right)$ for $n \in \mathbb{N}$.
Proof. In Lemma 3.1 it is shown that $\operatorname{dim}\left(H_{n}\right)=\operatorname{dim}\left(P_{n}\left(\mathbb{R}^{m-1}\right)\right)+\operatorname{dim}\left(P_{n-1}\left(\mathbb{R}^{m-1}\right)\right)$ which is also equal to the dimension of $P_{n}\left(\mathbb{R}^{m-1}\right) \times P_{n-1}\left(\mathbb{R}^{m-1}\right)$. We shall denote this space by $Z$ and we write the elements of $Z$ as $(f, g)$ for $f \in P_{n}\left(\mathbb{R}^{m-1}\right), g \in P_{n-1}\left(\mathbb{R}^{m-1}\right)$. Denote the monomials in $P_{n}\left(\mathbb{R}^{m-1}\right)$ by $x^{i}$ and the monomials in $P_{n-1}\left(\mathbb{R}^{m-1}\right)$ by $y^{j}$, then a basis of $Z$ is given by $\left\{\left(x^{i}, 0\right),\left(0, y^{j}\right)\right\}$, with $|i|=n$ or $|j|=n-1$.
By taking the $x_{m-1}$-dependence out of the multi-index, we can write $x^{i}=\prod_{a=1}^{m-1} x_{a}^{i_{a}} \equiv$ $x^{k} x_{m-1}^{l}$, where $k$ is a $(m-2)$-dimensional multi-index and $l=i_{m-2}$. We can use this to write the basis elements of $Z$ of the form $\left(x^{i}, 0\right)$ as $\left(x^{k} x_{m-1}^{l}, 0\right)=e_{k, l}$, which gives a one-to-one correspondence between those basis elements and elements of the set $\{(k, l) \in I, l \geq 0\}$.
We can also use relation to write the basis elements of $Z$ of the form $\left(0, y^{j}\right)$ as $\left(0, y^{k} y_{m-1}^{l-1}\right)=$ $e_{k,-l}, l<0$, which gives a one to one correspondence between basis elements and elements of the set $\{(k, l) \in I \mid l<0\}$.
By combining these two results, we see that $\operatorname{dim}\left(H_{n}\right)=\# I=\operatorname{dim}\left(\Phi_{n}\right)$.
Theorem 3.7. The functions $\phi_{n, k, l}$ defined in Definition 3.5 are polynomials in the coordinates $x_{i}$. We have the direct sum decomposition $\Phi_{n} \oplus r^{2} P_{n-2}=P_{n}$.

Proof. In the proof we will be using the trigonometric relations

$$
\begin{align*}
\sin (a+b) & =\sin (a) \cos (b)+\cos (a) \sin (b),  \tag{3.4}\\
\cos (a+b) & =\cos (a) \cos (b)-\sin (a) \sin (b) . \tag{3.5}
\end{align*}
$$

We can use these relations to write $\sin \left(l \theta_{m-1}\right)$ and $\cos \left(l \theta_{m-1}\right)$ as polynomials of degree $l$ in $\sin \left(\theta_{m-1}\right)$ and $\cos \left(\theta_{m-1}\right)$, which gives

$$
\begin{equation*}
\cos \left(l \theta_{m-1}\right)=\sum_{i=1}^{l} a_{i} \cos \left(\theta_{m-1}\right)^{l-i} \sin \left(\theta_{m-1}\right)^{i} \tag{3.6}
\end{equation*}
$$

and

$$
\begin{equation*}
\sin \left(l \theta_{m-1}\right)=\sum_{i=1}^{l} b_{i} \cos \left(\theta_{m-1}\right)^{l-i} \sin \left(\theta_{m-1}\right)^{i} . \tag{3.7}
\end{equation*}
$$

By using formulae (3.6) and (3.7), we can rewrite the functions $\phi_{n, k, l}$ in Lemma 3.5 as

$$
\phi_{n, k, l}= \begin{cases}r^{n} \prod_{i=1}^{m-1}\left[\sin \left(\theta_{i}\right)^{n-k_{1} \cdots-k_{i-1}} \cos \left(\theta_{i}\right)^{k_{i}}\right] \sum_{j=1}^{l} a_{i} \cos \left(\theta_{m-1}\right)^{l-j} \sin \left(\theta_{m-j}\right)^{i} ; & l \geq 0, \\ r^{n} \prod_{i=1}^{m-1}\left[\sin \left(\theta_{i}\right)^{n-k_{1} \cdots-k_{i-1}} \cos \left(\theta_{i}\right)^{k_{i}}\right] \sum_{j=1}^{-l} b_{i} \cos \left(\theta_{m-1}\right)^{-l-j} \sin \left(\theta_{m-1}\right)^{j} ; & l<0 .\end{cases}
$$

By comparing this with (3.3), we see that each of the terms in the sum equal to $\left(\zeta\left(r, \theta_{1} \ldots, \theta_{m-1}\right)^{\alpha}\right.$, for some multi-index $\alpha$, so each function $\phi_{n, k, l}\left(r, \theta_{1} \ldots, \theta_{m-1}\right)$ can be written as $p\left(\zeta\left(r, \theta_{1} \ldots, \theta_{m-1}\right)\right)$, for some polynomial $p$ on $\mathbb{R}^{m}$.
In the next step, we will use the decomposition

$$
P_{n}\left(\mathbb{R}^{m}\right)=\oplus_{i=0}^{n} P_{n-i}\left(\mathbb{R}^{m-2}\right) \otimes P_{i}\left(\mathbb{R}^{2}\right),
$$

naturally induces by the linear isomorphism $\mathbb{R}^{m} \rightarrow \mathbb{R}^{m} \oplus \mathbb{R}^{2}$ given by

$$
\left(x_{1}, \ldots, x_{m}\right) \rightarrow\left(x_{1}, \ldots x_{m-2}\right) \oplus\left(x_{m-1}, x_{m}\right) .
$$

Choose $i \geq 0$ arbitrary. Take as basis on $P_{n-i}\left(\mathbb{R}^{m-2}\right)$ the usual basis of monomials. For $p$ such a basis element define the subspace $S_{p}$ of $P_{n}\left(\mathbb{R}^{m}\right)$ by

$$
S_{p}=p P_{i}\left(\mathbb{R}^{2}\right) .
$$

Then

$$
P_{n}\left(\mathbb{R}^{m}\right)=\oplus_{i, p} S_{p} .
$$

By rewriting the elements of $S_{p}$ in spherical coordinates, we see that this space has a basis given by $e_{j}=q \cos ^{j}\left(\theta_{m-1}\right) \sin ^{i-j}\left(\theta_{m-1}\right)$, for $0 \leq j \leq i$, with

$$
q=p \times \prod_{j=1}^{m-2} \sin \left(\theta_{j}\right)^{i}=r^{n} \prod_{j=1}^{m-2}\left(\sin \left(\theta_{j}\right)^{n-k_{1}-\cdots-k_{j}} \cos \left(\theta_{j}\right)^{k_{j}}\right)
$$

The values $k_{j}$ are fixed by the choice of $p$ and we can see that

$$
\Phi_{n} \cap S_{p}=\left\{\phi_{n, k, \pm l}\right\} .
$$

Note that for $i=0$, the space $S_{p}$ is 1-dimensional space with basis element $q$. In this formula $q$ contains all dependencies besides the $\theta_{m-1}$-dependence. The elements of $S_{p}$ only differ in the $\theta_{m-1}$-dependence.
Another basis of $S_{p}$ is given by

$$
f_{j}= \begin{cases}q \cos \left(j \theta_{m-1}\right) & \text { if } j \text { is even, }  \tag{3.8}\\ q \sin \left((j-1) \theta_{m-1}\right) & \text { if } j \text { is odd. }\end{cases}
$$

As before $0 \leq j \leq i$. By using the indices $n, l=n-i$ and the indices $k_{j}$ from the definition of $p$, we see that $f_{j}=\phi_{n, k, l}$ for even $j$, and $f_{j}=\phi_{n, k,-l}$ for odd $j$. The other elements are clearly in $r^{2} P_{n-2}\left(\mathbb{R}^{m}\right)$, which shows that $S_{p}$ can be decomposed as

$$
\begin{equation*}
S_{p}=\left(S_{p} \cap r^{2} P_{n-2}\right) \oplus\left(S_{p} \cap \Phi_{n}\right) . \tag{3.9}
\end{equation*}
$$

For $i=0$, the space $S_{p}$ is one-dimensional and has $f_{0}=q$ as only basis element. The basis element $q=\phi_{n, k, 0}$ for the $n, k$ associated with $p$.
The elements $f_{j}$ defined in (3.8) depend on the choice of the basis element $p$. The set of all elements $f_{j}(p)$ is a basis for all of $P_{n}\left(R^{m}\right)$, because

$$
P_{n}\left(R^{m}\right)=\oplus_{p, i} S_{p}
$$

Together with (3.9), this shows that we have the decomposition

$$
P_{n}\left(R^{m}\right)=X \oplus \Phi_{n},
$$

where $X=r^{2} P_{n-2} \cap \operatorname{span}\left(f_{j}(p)\right)$. Because of Lemma 3.6 the space $X=P_{n} / \Phi_{n}$ is a linear space of the same dimension as $r^{2} P_{n-2}$, which shows that $X=r^{2} P_{n-2}$ and

$$
P_{n}\left(R^{m}\right)=r^{2} P_{n-2} \oplus \Phi_{n} .
$$

Now we have enough tools to construct a basis of $H_{n}$.
Consider the two direct sum decompositions of $P_{n}$, given by

$$
P_{n}=H_{n} \oplus r^{2} P_{n-2} \text { and } P_{n}=\Phi_{n} \oplus r^{2} P_{n-2}
$$

The associated inclusion maps are given by $f_{1}: H_{n} \rightarrow P_{n}$ and $f_{2}: \Phi_{n} \rightarrow P_{n}$. The associated projections are given by $\pi_{1}: P_{n} \rightarrow H_{n}$ and $\pi_{2}: P_{n} \rightarrow \Phi_{n}$.
The map $\pi_{1} \circ f_{2}$ is a bijective linear map from $\Phi_{n}$ into $H_{n}$, with inverse $\pi_{2} \circ f_{1}$, since the equivalence classes of $P_{n}$ with respect to $\pi_{1}$ and $\pi_{2}$ are the same. This means in particular, that a basis of $\Phi_{n}$ is sent to a basis of $H_{n}$, so one basis of $H_{n}$ is given by

$$
\psi_{n, k, l}=\pi_{2} \circ i_{1}\left(\phi_{n, k, l}\right)=\pi_{2}\left(\phi_{n, k, l}\right),
$$

where the elements $\phi_{n, k, l}$ were defined in Definition 3.5.
Let $f$ be an element of $P_{n}$. By Theorem 2.1, we can write $f=\sum|x|^{2 j} f_{n-2 j}$, where each $f_{n-2 j} \in H_{n-2 j}$ and each $f_{n-2 j}$ is unique. We want to find linear maps $\pi_{n j}: P_{n} \rightarrow H_{n-2 j}$ such that $\pi_{n j} f=f_{n-2 j}$. For this we need that

$$
|x|^{2 i} \Delta^{i} f=\sum \lambda_{i j}|x|^{2 j} f_{n-2 j}, \quad(0 \leq i \leq n / 2),
$$

where the constants $\lambda_{i j}$ can be found from Corollary 2.4. This gives us a linear system of $1+\lfloor n / 2\rfloor$ equations, in the $1+\lfloor n / 2\rfloor$ unknowns $f_{m-2 j}$. We will look again at the constants $\lambda_{i j}$ in Theorem 6.26 on page 38, and we will solve this system of equations in Corollary 6.30. The linear function $\pi_{n 0}: P_{n} \rightarrow H_{n}$ is equal to the projection $\pi_{2}$, which was used to construct the basis of $H_{n}$.

## Chapter 4

## A more general description of Fischer decompositions

We can look at the Fischer decomposition in a more abstract way, which will give us an easier way to prove existence of such decompositions.

Definition 4.1. Let $V$ be a vector space, with inner product $\langle\cdot, \cdot\rangle$. Let $A, B$ be linear maps from $V$ to $V$. Then $A, B$ are formal adjoints if and only if $\langle A f, g\rangle=\langle f, B g\rangle$, for all $f, g \in V$.

Lemma 4.2. If $B$ is a formal adjoint of $A: V \rightarrow W$ then $B$ is unique.
Proof. Suppose $\tilde{B}$ is another formal adjoint of $A$. Then for all $f \in V$ and all $g \in W$ we have that $\langle f, B g\rangle=\langle A f, g\rangle=\langle f, \tilde{B} g\rangle$, so $B g=\tilde{B} g$ for all $g \in W$.

Lemma 4.3. Let $V$ be a graded vector space $V=\oplus_{n \in \mathbb{Z}} V_{n}$, with $\operatorname{dim} V_{n}<\infty \forall n \in \mathbb{Z}$ and inner product $\langle\cdot, \cdot\rangle$, such that $V_{n} \perp V_{m}$ if $n \neq m$. If the map $A: V \rightarrow V$ has degree $k$ with respect to this grading, which means that $A\left(V_{n}\right) \subset V_{n+k}$, then $A$ has a formal adjoint.
Proof. Write $A_{n}$ for the map $A \mid V_{n}: V_{n} \rightarrow V_{n+k}$. The spaces $V_{n}$ and $V_{n+k}$ are finite dimensional, so for $w \in V_{n+k}$, we can define the vector $A_{n}^{*}(w) \in V_{n}$ by $\left\langle A_{n}^{*}(w), \cdot\right\rangle=\langle w, A(\cdot)\rangle$ for all $v \in V_{n}$. The functionals $\left\langle A_{n}^{*}(w), \cdot\right\rangle$ and $\langle w, A(\cdot)\rangle$ are elements of $V_{n}^{*}$, the dual space of $V_{n}$. The map $A_{n}^{*}: V_{n+k} \rightarrow V_{n}$ is the adjoint of $A_{n}$. Let $B=\oplus_{n \in \mathbb{Z}} A_{n}^{*}$, then $B$ is the formal adjoint of $A$.

Remark 4.4. In literature the adjoint of the operator $A$ is often denoted by $A^{*}$.
Lemma 4.5. Let $V=\oplus_{n \in \mathbb{Z}} V_{n}$ be a graded vector space with inner product $\langle\cdot, \cdot\rangle$, such that $V_{n} \perp V_{m}$ if $n \neq m, V_{n}=0$ if $n<0$ and $V_{n}$ is finite dimensional for all $n \in \mathbb{Z}$.
Let $A$ be a linear map from $V$ to $V$ of degree $-k$, so $A\left(V_{n}\right) \subseteq V_{n-k}$. Let $B$ be the formal adjoint of $A$. Then the map $B$ is a linear map of degree $k$.

Proof. Let $v \in V_{s}$ and $w \in V_{t}$. Then $\langle B v, w\rangle=\langle v, A w\rangle=0$ if $s \neq t-k$.
So for all $t \neq s+k$ and for all $w \in V_{t},\langle b(v), w\rangle=0$, which implies that $b(v) \in V_{s+k}$.
Lemma 4.6. Let $V, W$ be finite dimensional linear spaces with positive definite inner product, $A: V \rightarrow W$ and $B: W \rightarrow V$ linear maps and let $A$ be the adjoint of $B$. Then we have that $V=\operatorname{im}(B) \oplus \operatorname{ker}(A)$. We also have that $W=\operatorname{im}(A) \oplus \operatorname{ker}(B)$.

Proof. For the first part need to show that $\operatorname{ker}(A)$ and $\operatorname{im}(B)$ are orthogonal, that their intersection is 0 . We also need to show that no nonzero element of $V$ is orthogonal to both $\operatorname{ker}(A)$ and $\operatorname{im}(B)$. Let $x \in \operatorname{ker}(A)$. Then $\langle A x, y\rangle=0$ for all $y \in W$. From this follows that $\langle x, B y\rangle=0, \forall y \in W$ and $x \perp \operatorname{im}(B)$.
Let $x \in \operatorname{im}(B)$. Then $x=B z$ for some $z \in W$. From this follows that $\langle x, B z\rangle=\langle A x, z\rangle \neq 0$ by the positive definiteness of the inner product, so $x \notin \operatorname{ker}(A)$.
Let $x \in \operatorname{im}(B)$ and $y \in \operatorname{ker}(A)$. Then $\langle x, y\rangle=\langle B z, y\rangle=\langle z, A y\rangle=0$, so $x \perp \operatorname{ker}(A)$.
Let $x \perp \operatorname{im}(B)$ and $x \perp \operatorname{ker}(A)$. Then $\langle A x, y\rangle=\langle x, B y\rangle=0, \forall y \in W$, which implies $x \in \operatorname{ker}(A)$, so $x=0$.
This means we can write $V=\operatorname{ker}(A) \oplus \operatorname{im}(B)$, because $\operatorname{ker}(A) \perp \operatorname{im}(B), \operatorname{ker}(A) \cap \operatorname{im}(B)=0$ and no nonzero element of $V$ is orthogonal to both $\operatorname{ker}(A)$ and $\operatorname{im}(B)$.
The second part follows by interchanging the roles of $A$ and $B$.
Corollary 4.7. Let $V, W$ be finite dimensional linear spaces with positive definite inner product, $A: V \rightarrow W$ and $B: W \rightarrow V$ linear maps and let $A$ be the adjoint of $B$. Then $A$ is surjective if and only if $B$ is injective.

Proof. If $A$ is surjective, we have that $W=\operatorname{im}(A)$, so by Lemma 4.6 the Kernel of $B$ is 0 and $B$ is injective.
If $B$ is injective, it follows from Lemma 4.6 that $W=\operatorname{im}(A)$, so $A$ is surjective.
Theorem 4.8. Let $V=\oplus_{n \in \mathbb{Z}} V_{n}$ be a graded vector space with inner product $\langle\cdot, \cdot\rangle$, such that $V_{n} \perp V_{m}$ if $n \neq m, V_{n}=0$ if $n<0$ and $V_{n}$ is finite dimensional for all $n \in \mathbb{Z}$.
Let $A$ be a surjective linear map from $V$ to $V$ of degree $-k$ with formal adjoint $B$. Define $H_{n}=V_{n} \cap \operatorname{ker}(A)$. Then the spaces $V_{n}$ can be decomposed as

$$
V_{n}=\bigoplus_{i=0}^{\lfloor n / k\rfloor} B^{i}\left(H_{n-k i}\right)
$$

This decomposition is a Fischer decomposition.
Proof. By Lemma 4.5 the space $B\left(V_{n-k}\right) \subseteq V_{n}, \forall n \in \mathbb{Z}$. By the surjectivity of $A$, we have that $A\left(V_{n}\right)=V_{n-k}$. By Corollary 4.7 it follows that $\left.B\right|_{V_{n-k}}$ is injective for all $n \in \mathbb{Z}$.
By Lemma 4.6 we have the decomposition

$$
V_{n}=B\left(V_{n-k}\right) \oplus V_{n} \cap \operatorname{ker}(A)=B\left(V_{n-k}\right) \oplus H_{n} .
$$

By repeating this argument we find that

$$
V_{n}=B\left(B\left(V_{n-2 k}\right) \oplus H_{n-k}\right) \oplus H_{n}=B^{2}\left(V_{n-2 k}\right) \oplus B\left(H_{n-k}\right) \oplus H_{n},
$$

where we the injectivity of $B$ is needed to show that sum on the right hand side is a direct sum. Since $V_{i}=0$ for $i<0$, we only have to repeat these steps a finite number of times, which leads to

$$
V_{n}=\bigoplus_{i}^{\lfloor n / k\rfloor} B^{i}\left(H_{n-k i}\right) .
$$

## Chapter 5

## Fischer decompositions of $P\left(\mathbb{R}^{m}\right)$

Let $P=P\left(\mathbb{R}^{m}\right)$ and let $P_{n}$ be the space of homogeneous polynomials of degree $n$. Let $p \in P_{k}, q \in P_{n}$. It was shown by E. Fischer in [16] that $q=a p+b$, with $a \in P_{n-k}$ and $b \in P_{n} \cap \operatorname{ker}(p(\partial))$. Here $p(\partial)$ is the element of the ring $\mathbb{R}\left[\partial_{1}, \ldots \partial_{m}\right]$, which is obtained from $p(x)$ by replacing each instance of $x_{j}$ with $\partial_{j}$. This will be made more precise in Definition 5.1. Repeated use of the mentioned result leads to the Fischer decomposition

$$
P=\bigoplus_{n=0}^{\infty} \bigoplus_{i=0}^{\lfloor n / k\rfloor} p^{i}\left(P_{n-k i} \cap \operatorname{ker}(p(\partial))\right.
$$

In this chapter we will use the results from Chapter 4 to show the existence of this Fischer decomposition in another way.
Before we do this, we need to construct an appropriate inner product on $P$. A special case of this is the harmonic Fischer decomposition, which was used in Section 2. However, the proof with the $\mathfrak{s l}_{2}$-representation does not work in the general case.
Definition 5.1. Let $p$ be a formal power series in $m$ variables, $x_{1}, \ldots, x_{m}$.
Define by $p(\partial)$ the formal power series, which is obtained by replacing the variable $x_{i}$ with the partial derivative $\partial / \partial x_{i}$ in the expression of $p(x)$.
We will sometimes use the notation $p\left(\partial_{x}\right)$ to emphasize that we take partial derivatives with respect to the variables $x_{1}, \ldots, x_{m}$.
The operator $p(\partial)$ is an element of the ring $R=\mathbb{R}\left[\left[\partial_{1}, \ldots, \partial_{m}\right]\right]$. We have the natural action of $R$ on $P\left(\mathbb{R}^{m}\right)$ given by

$$
(r, p) \mapsto r(\partial) p(x), r \in R, p \in P
$$

By using multi-index notation, we can write

$$
r=\sum_{n=0}^{\infty} \sum_{|\alpha|=n} c_{\alpha} \partial^{\alpha}, \text { with } c_{\alpha} \in \mathbb{R}
$$

Let $p$ be a polynomial of degree at most $k$, then

$$
r(\partial) p=\sum_{n=0}^{\infty} \sum_{|\alpha|=n} c_{\alpha} \partial^{\alpha} p=\sum_{n=0}^{k} \sum_{|\alpha|=n} c_{\alpha} \partial^{\alpha} p
$$

which is a finite sum, so the action is well-defined for each element of $R$.

Definition 5.2. For $p, q \in P=P\left(\mathbb{R}^{m}\right)$, define the bilinear form $[\cdot, \cdot]: P \times P \rightarrow \mathbb{R}$ by

$$
[p, q]=\left.p(\partial) q(x)\right|_{x=0} .
$$

Lemma 5.3. The form $[\cdot, \cdot]$ defines an inner product on $P$. On the monomials this form is given by $\left[x^{k}, x^{l}\right]=\delta_{k l} k!$, where we have used the multi-index notation.

Proof. This form is clearly bilinear, so we can use its definition on the monomials to prove that it is an inner product. Note that in the one dimensional case, without using the multi-index notation, we have

$$
\left\langle x^{k}, x^{l}\right\rangle=\left.\frac{\partial^{k}}{\partial x^{k}} x^{l}\right|_{x=0}=\delta_{k l} k!
$$

Since the partial derivatives commute, we find in the multidimensional case that

$$
\begin{aligned}
\left\langle x^{k}, x^{l}\right\rangle & =\prod_{i=1}^{m}\left(\frac{\partial^{k_{i}}}{\partial x_{1}^{k_{i}}} x_{i}^{l_{i}}\right)_{x=0} \\
& =\prod_{i=1}^{m}\left(\delta_{k_{i} l_{i}} k_{i}!\right) \\
& =\delta_{k l} k!
\end{aligned}
$$

where we have used multi-index notation.
Choose $p, q \in P$ arbitrary. We can write these polynomials as sums of monomials, which leads to

$$
[p, q]=\sum_{k, l} p_{k} q_{l}\left[x^{k}, x^{l}\right]=\sum_{k} k!p_{k} q_{k} .
$$

From this formula we also see that the product is symmetric and that $[p, p]=0$ implies that $p=0$, so $[\cdot, \cdot]$ is an inner product on $P$.
Definition 5.4. The reproducing kernel $\hat{K}: \mathbb{R}^{n} \times \mathbb{R}^{n} \rightarrow \mathbb{R}$ is defined by

$$
\hat{K}(x, y)=\exp \langle x, y\rangle .
$$

We also define $\hat{K}_{n}(x, y): \mathbb{R}^{n} \times \mathbb{R}^{n} \rightarrow \mathbb{R}$ by $\hat{K}_{n}(x, y)=\langle x, y\rangle^{n} / n!$.
Note that each $\hat{K}_{n}$ is a homogeneous polynomial of degree $n$ in the variables $x_{i}$, with the variables $y_{i}$ viewed as parameters. Each $\hat{K}_{n}$ is also a homogeneous polynomial of degree $n$ in the variables $y_{i}$, with the variables $x_{i}$ viewed as parameters. We also have that $\hat{K}(x, y)=$ $\sum_{n=0}^{\infty} \hat{K}_{n}(x, y)$.
By using Definition 5.1 twice, we can view $\hat{K}\left(\partial_{x}, \partial_{y}\right)$ as an element of the polynomial ring $R \times R$, where the first component contains the $\partial_{x^{\prime}}$-terms and the second component the $\partial_{y^{-}}$ terms.
Lemma 5.5. Let $p$ be a polynomial in $m$ variables. By Definition 5.1, we can view $\hat{K}\left(\partial_{x}, y\right)$ as an element of $R$, with the $y_{i}$ as parameters. Then we have

$$
\hat{K}\left(\partial_{x}, y\right) p(x)=p(y) .
$$

By using the natural action of $R \times R$ on $P \times P$ it follows that

$$
\left.\hat{K}\left(\partial_{x}, \partial_{y}\right) p(x) q(y)\right|_{y=0}=[p, q] .
$$

Proof. In the following we will use the multi-index notation. Since $\hat{K}(x, y)$ has a convergent series expansion, we can write

$$
\hat{K}(x, y)=\sum_{n=0}^{\infty} \sum_{|\alpha|=n} c_{\alpha} x^{\alpha} y^{\alpha}
$$

and by using Definition 5.1, we get

$$
\hat{K}\left(\partial_{x}, y\right)=\sum_{\alpha} c_{\alpha} \partial_{x}^{\alpha} y^{\alpha}
$$

with constants $c_{\alpha} \in \mathbb{R}$. Here we have an infinite sum over all possible values of $\alpha$, but only a finite number of terms is nonzero, so the sum is well-defined.
Because of linearity, we only have to check the lemma for monomials.

$$
\begin{align*}
\hat{K}\left(\partial_{x}, y\right) x^{\beta} & =\sum_{\alpha} c_{\alpha} \partial_{x}^{\alpha} y^{\alpha} x^{\beta} \\
& =\sum_{\alpha} c_{\alpha} y^{\alpha}\left[x^{\alpha}, x^{\beta}\right] \\
& =\sum_{\alpha, \beta} c_{\alpha} \alpha!\delta_{\alpha \beta} y^{\alpha} \\
& =\beta!c_{\beta} y^{\beta} \tag{5.1}
\end{align*}
$$

where we have used the inner product defined in Lemma 5.3.
The constant $c_{\beta}$ is the coefficient of $x^{\beta} y^{\beta}$ in $\hat{K}(x, y)$. Assume $|\beta|=k$. We only have to expand $\hat{K}_{k}(x, y)$ to find the coefficient $c_{\beta}$.
For this we need the multinomial coefficients given by

$$
\begin{equation*}
\frac{1}{k!}\left(x_{1}+\cdots+x_{m}\right)^{k}=\sum_{\alpha=|k|} \frac{1}{\alpha!} x^{\alpha} \tag{5.2}
\end{equation*}
$$

For $m=2$ this formula gives the binomial coefficients

$$
\frac{1}{k!}\left(x_{1}+x_{2}\right)=\sum_{j=0}^{k} \frac{1}{j!} \frac{1}{(k-j)!} x_{1}^{j} x_{2}^{k-j}
$$

The general case can be proven by induction over $m$, so suppose the formula is correct for $m=l$, then by using the binomial coefficients

$$
\begin{aligned}
\frac{1}{k!}\left(\left(x_{1}+\cdots+x_{l}\right)+x_{l+1}\right)^{k} & =\sum_{j=0}^{k} \frac{1}{j!} \frac{1}{k-j}\left(x_{1}+x_{l}\right)^{j} x_{l+1}^{k-j} \\
& =\sum_{j=0}^{k}\left(\frac{1}{(k-j)!} x_{l+1}^{k-j} \sum_{\alpha=|j|} \frac{1}{\alpha!} x^{\alpha}\right) \\
& =\sum_{\beta=|k|} \frac{1}{\beta!} x^{\beta},
\end{aligned}
$$

where $\alpha$ is a multi-index over $\mathbb{N}^{l}$ and $\beta$ is a multi-index over $\mathbb{N}^{l+1}$.
Since $K_{k}(x, y)=\left(x_{1} y_{1}+\cdots+x_{m} y_{m}\right)^{k} / k$ !, by equation (5.2) the coefficients in equation (5.1), are given by $c_{\beta}=1 / \beta$ ! and so

$$
\hat{K}\left(\partial_{x}, y\right) x^{\beta}=y^{\beta},
$$

and by linearity

$$
\hat{K}\left(\partial_{x}, y\right) p(x)=p(y)
$$

By the same argument, we see that

$$
\hat{K}\left(\partial_{x}, \partial_{y}\right) p(x)=p\left(\partial_{y}\right)
$$

and

$$
\left.\hat{K}\left(\partial_{x}, \partial_{y}\right) p(x) q(y)\right|_{y=0}=\left.p\left(\partial_{y}\right) q(y)\right|_{y=0}=[p, q] .
$$

The first formula in Lemma 5.5 is the reason why $\hat{K}$ is called a reproducing kernel. For the operator $\hat{K}_{n}$ we have the following property

Lemma 5.6. Define by $\pi_{n}$ the projection from $P$ onto $P_{n}$. Then

$$
\left.K_{n}\left(\partial_{x}, y\right) p(x)\right|_{x=0}=\pi_{n} p(y) .
$$

Proof. Take $p \in P_{k}$. Assume $k>n$, then $K_{n}\left(\partial_{x}, y\right) p(x)$ is a homogeneous polynomial of degree $k-n$ in the variables $x$. Since $k-n>0$ we have that

$$
\left.K_{n}\left(\partial_{x}, y\right) p(x)\right|_{x=0}=0=\pi_{n} p(y),
$$

because $P_{k} \cap P_{n}=0$, for $n \neq k$.
Next assume $k<n$. Then $K_{n}\left(\partial_{x}, y\right) p(x)=0$, since we are differentiating a polynomial of degree $k$, more than $k$ times, which gives 0 . So again

$$
\left.K_{n}\left(\partial_{x}, y\right) p(x)\right|_{x=0}=0=\pi_{n} p(y) .
$$

For $n=k$, we have by Lemma 5.5 and the previous two statements

$$
p(y)=\hat{K}\left(\partial_{x}, y\right) p(x) \sum_{n=0}^{\infty} \hat{K}_{n}\left(\partial_{x}, y\right) p(x)=\hat{K}_{k}\left(\partial_{x}, y\right) p(x)=\pi_{k} p(x) .
$$

The kernel $\hat{K}$ extended to a complex differentiable map $\mathbb{C} \times \mathbb{C} \rightarrow \mathbb{C}$ is also used to define the Fourier transformation, which is given by

$$
\mathscr{F}(f)(y)=\frac{1}{(2 \pi)^{m / 2}} \int_{\mathbb{R}^{m}} f(x) \hat{K}(x,-i y) d x=\frac{1}{(2 \pi)^{m / 2}} \int_{\mathbb{R}^{m}} f(x) \exp (-i\langle x, y\rangle) d x,
$$

for $f \in L^{2}\left(\mathbb{R}^{m}\right)$.
Theorem 5.7. Let $p=\sum_{|\alpha|=k} a_{\alpha} x^{\alpha},\left(a_{\alpha} \in \mathbb{R}\right)$ be an arbitrary homogeneous polynomial of degree $k$ and let $D$ be the associated differential operator $p(\partial / \partial x)$. Define the subspaces $H_{n}=P_{n} \cap \operatorname{ker}(D)$. Then $P_{n}$ admits the Fischer decomposition

$$
P_{n}=\bigoplus_{i} p^{i} H_{n-k i} .
$$

Proof. Take the vector space $V=P$, with the inner product defined in Lemma 5.3. The space $P$ has a natural grading given by $V_{n}=P_{n}$. By a direct computation we see that $p V_{n} \subset V_{n+k}$ and $D V_{n} \subset V_{k-n}$.
We also have that

$$
\begin{aligned}
{[p f, g] } & =\sum_{|\alpha|=k} a_{\alpha}(\partial)^{\alpha} f(\partial) g \\
& =\sum_{\alpha} a_{\alpha} f(\partial)(\partial)^{\alpha} g \\
& =[f, D g]
\end{aligned}
$$

so $p$ and $D$ are formal adjoints.
Since $p: P_{l} \rightarrow P_{k+l}, l \in \mathbb{N}$ is injective, it follows by Corollary 4.7 that $D: P_{k+l} \rightarrow P_{l}$ is surjective, from which it follows that $D: P \rightarrow P$ is surjective.
Now we can apply Theorem 4.8 and we obtain the Fischer decomposition.

## Chapter 6

## Dunkl operators

The Dunkl operator $T_{u},\left(u \in \mathbb{R}^{m}\right)$ is a generalization of the partial derivative $\partial_{u}$, which is still homogeneous of degree -1 . In the following chapters we will define this operator, and show that there are related decompositions similar to the Fischer decompositions in Theorem 2.1 and Theorem 5.7. We will also show that there is an equivalent of the Fourier transform for the Dunkl operators and use this transform to solve certain types of differential-difference equations.

We start from the space $\mathbb{R}^{m}$, with the usual inner product. For $\alpha \in \mathbb{R}^{m} \backslash\{0\}$ the reflection $r_{\alpha}$ is defined by

$$
r_{\alpha}(x)=x-\frac{2\langle\alpha, x\rangle}{\langle\alpha, \alpha\rangle} \alpha,\left(x \in \mathbb{R}^{m}\right) .
$$

Definition 6.1. A root system in $\mathbb{R}^{m}$ is a finite subset $R$ of $\mathbb{R}^{m} \backslash\{0\}$, such that $r_{\alpha}(R)=R$, for all $\alpha \in R$.
A root system is called reduced if $R \cap \mathbb{R} \alpha= \pm \alpha, \forall \alpha \in R$.
From now on, we assume that $R$ is reduced. The root system can be written as a disjoint union $R=R_{+} \cup-\left(R_{+}\right)$, where the two sets are separated by a hyperplane through 0 . We can renormalize all the roots such that $\langle\alpha, \alpha\rangle=2$.
Definition 6.2. The Weyl group of $R$ is the finite group $G$ which is generated by the reflections $r_{\alpha}$, for $\alpha \in R$. The Weyl group has a natural action on $R$, which is given by $(w, \alpha) \mapsto w \alpha$, for $w \in G, \alpha \in R$. A weight function on $R$ is a $G$-invariant function $k: R \rightarrow \mathbb{R}$.
It is convenient to use the notation $k(\alpha)=k_{\alpha}, \alpha \in R$. In the following chapter, we will assume that $k_{\alpha}$ is positive.
We will shortly discuss negative weight functions in Section 7.3 .
Definition 6.3. [9, Def. 1.3, 1.4] For a given root system $R$ with weight function $k$, the $k$-gradient, or Dunkl gradient, $\nabla_{k}: C^{1}\left(\mathbb{R}^{m}\right) \rightarrow C\left(\mathbb{R}^{m}\right) \otimes \mathbb{R}^{m}$ is the operator defined by

$$
\nabla_{k} f(x)=\nabla f(x)+\sum_{\alpha \in R^{+}} \alpha k_{\alpha} \frac{f(x)-f\left(r_{\alpha} x\right)}{\langle\alpha, x\rangle}=\nabla f(x)+\frac{1}{2} \sum_{\alpha \in R} \alpha k_{\alpha} \frac{f(x)-f\left(r_{\alpha} x\right)}{\langle\alpha, x\rangle} .
$$

For a nonzero vector $u \in \mathbb{R}^{m}$, the associated Dunkl operator $T_{u}: C^{1}\left(\mathbb{R}^{m}\right) \rightarrow C\left(\mathbb{R}^{m}\right)$, is defined by

$$
\left(T_{u} f\right)(x)=\left\langle\nabla_{k} f(x), u\right\rangle
$$

To simplify the notation we define the difference operator $s_{\alpha}: C^{1}\left(\mathbb{R}^{m}\right) \rightarrow C^{0}\left(\mathbb{R}^{m}\right)$ by

$$
s_{\alpha} f=\frac{f(x)-f\left(r_{\alpha} x\right)}{\langle\alpha, x\rangle},
$$

so the Dunkl gradient can be written as

$$
\nabla_{k} f=\nabla f+\sum_{\alpha \in R_{+}} \alpha k_{\alpha} s_{\alpha} f .
$$

Note that $\nabla_{k}(f)$ is not well-defined at the points $x \in \mathbb{R}^{m}$, with $\langle x, \alpha\rangle=0$ for some $\alpha \in R$, because $s_{\alpha} f(x)=(f(x)-f(x)) / 0=0 / 0$ at these points. However, if $f \in C^{1}\left(\mathbb{R}^{m}\right)$, we can use the identity

$$
\begin{equation*}
f(x)-f\left(r_{\alpha} x\right)=\langle x, \alpha\rangle \int_{0}^{1} \partial_{\alpha} f\left(t x+(1-t) r_{\alpha} x\right) d t \tag{6.1}
\end{equation*}
$$

to define the Dunkl operators for all $x \in \mathbb{R}^{m}$, because the directional derivative is continuous (See [14, p. 3]).
This identity is proven by noting that

$$
\frac{d}{d t} f\left(t x+(1-t) r_{\alpha} x\right)=\sum_{i=1}^{n} \frac{\partial}{\partial x_{i}}\left(f\left(t x+(1-t) r_{\alpha} x\right) \alpha_{i}\right)\langle x, \alpha\rangle=\partial_{\alpha} f\left(t x+(1-t) r_{\alpha} x\right)\langle x, \alpha\rangle .
$$

By using (6.1) we see that

$$
s_{\alpha} f(x)=\int_{0}^{1} \partial_{\alpha} f\left(t x+(1-t) r_{\alpha} x\right) d t \text { for }\langle x, \alpha\rangle \neq 0
$$

Thus, if $f \in C^{1}\left(\mathbb{R}^{m}\right)$, it follows that $s_{\alpha} f$ uniquely extends to a continuous function on $\mathbb{R}^{m}$. By using this, the functions $T_{u} f, f \in C^{1}\left(\mathbb{R}^{m}\right)$ given in Definition 6.3 are extended to continuous functions on all of $\mathbb{R}^{m}$.

Lemma 6.4. The Dunkl operators send homogeneous polynomials of degree $n$ to homogeneous polynomials of degree $n-1$.

Proof. It is clear that the partial derivative sends homogeneous polynomials of degree $n$ to homogeneous polynomials of degree $n-1$, so we only need to show that the operators $s_{\alpha}, \alpha \in R_{+}$, have this property as well. It suffices to show this for a monomial $p(x)=x^{\beta}$ of degree $n$. This gives

$$
\begin{aligned}
s_{\alpha} p(x) & =\frac{p(x)-p(x-\langle x, \alpha\rangle \alpha)}{\langle x, \alpha\rangle} \\
& =\frac{x^{\beta}-(x-\langle x, \alpha\rangle \alpha)^{\beta}}{\langle x, \alpha\rangle} \\
& =\frac{\left(x^{\beta}-x^{\beta}+\langle x, \alpha\rangle p_{\beta}(x)\right)}{\langle x, \alpha\rangle} \\
& =p_{\beta}(x),
\end{aligned}
$$

where $p_{\beta}$ is a homogeneous polynomial of degree $n-1$ for each $\beta$. So $T_{u} p_{n}$ is also a homogeneous polynomial of degree $n-1$.

Lemma 6.5. [14, Thm 2.4, Cor 2.5] For $f, g \in C^{1}\left(\mathbb{R}^{m}\right)$, the Leibniz rule can be generalized to

$$
T_{u}(f g)(x)=\left(T_{u} f(x)\right) g(x)+f(x) T_{u} g(x)-\sum_{\alpha \in R_{+}} k_{\alpha} \frac{\langle\alpha, u\rangle}{\langle\alpha, x\rangle}\left(f(x)-f\left(r_{\alpha}(x)\right)\left(g(x)-g\left(r_{\alpha}(x)\right) .\right.\right.
$$

If $f$ or $g$ is $G$-invariant this simplifies to

$$
T_{u}(f g)=T_{u}(f) g+f T_{u} g
$$

Proof. A direct computation shows that

$$
\begin{aligned}
T_{u}(f g)(x) & =\left(\partial_{u} f\right)(x) g(x)+f(x)\left(\partial_{u} g\right)(x)+\sum_{\alpha \in R_{+}} k_{\alpha} \frac{\langle\alpha, u\rangle}{\langle\alpha, x\rangle}\left(f(x) g(x)-f\left(r_{\alpha}(x)\right) g\left(r_{\alpha}(x)\right)\right. \\
& =\left(\partial_{u} f\right)(x) g(x)+f(x)\left(\partial_{u} g\right)(x)+\sum_{\alpha \in R_{+}} k_{\alpha} \frac{\langle\alpha, u\rangle}{\langle\alpha, x\rangle}\left[f(x) g(x)-f\left(r_{\alpha}(x)\right) g\left(r_{\alpha}(x)\right)\right. \\
& \left.+f(x) g\left(r_{\alpha}(x)\right)-f(x) g\left(r_{\alpha}(x)\right)+f\left(r_{\alpha} x\right) g(x)-f\left(r_{\alpha}(x)\right) g(x)+f(x) g(x)-f(x) g(x)\right] \\
& =T_{u} f(x) g(x)+f(x) T_{u} g(x)-\sum_{\alpha \in R_{+}} k_{\alpha} \frac{\langle\alpha, u\rangle}{\langle\alpha, x\rangle}\left(f(x)-f\left(r_{\alpha}(x)\right)\left(g(x)-g\left(r_{\alpha}(x)\right) .\right.\right.
\end{aligned}
$$

In the last step was used that first and fourth term sum to the difference part of $T_{u} f(g)$ and that the sixth and the seventh term sum to the difference part of $f T_{u}(g)$.
If $f$ is $G$-invariant, $f(x)=f\left(r_{\alpha}(x)\right) \forall \alpha$, so the third term is zero. Also, if $g$ is $G$-invariant, $g(x)=g\left(r_{\alpha}(x)\right) \forall \alpha$, so the third term is again zero.
The left regular action of $G$ on the space $C\left(\mathbb{R}^{m}\right)$ is given by $L(w) f(x)=f\left(w^{-1} x\right)$, for $w \in G$ and $f \in \mathbb{R}$.
Lemma 6.6. [9, p. 169] For $f \in C^{1}\left(\mathbb{R}^{m}\right), w \in G$, we have the relation

$$
\left(\nabla_{k} L\left(w^{-1}\right) f\right)(x)=w \nabla_{k} f(w x) .
$$

Proof. For $g \in C^{1}\left(\mathbb{R}^{m}\right)$ we have that

$$
\begin{aligned}
\nabla_{k}\left(L\left(w^{-1}\right) f\right)(x) & =\sum_{i=1, j=1}^{m} e_{i}\left(\frac{\partial}{\partial x_{i}} f\right)(w x) w_{i j}+\sum_{i=1}^{m} \sum_{\alpha \in R_{+}} k_{\alpha} \alpha_{i} \frac{f(w x)-f\left(r_{\alpha} w x\right)}{\langle w x, \alpha\rangle} \\
& =w(\nabla f)(w x)+\sum_{i=1}^{m} \sum_{\alpha \in R_{+}} k_{\alpha} \alpha_{i} \frac{f(w x)-f\left(r_{\alpha} w x\right)}{\langle w x, \alpha\rangle}
\end{aligned}
$$

If $w r_{\beta}=r_{\alpha} w$, we have that $k_{\alpha}=k_{\beta}$ and $w r_{\beta} w^{-1} \alpha=r_{\alpha} \alpha=-\alpha$, so $r_{\beta} w^{-1} \alpha=-w^{-1} \alpha$, which shows that $\beta=w^{-1} \alpha$. Applying this to the equation gives

$$
\begin{aligned}
\nabla_{k}\left(L\left(w^{-1}\right) f\right)(x) & =w(\nabla f)(w x)+\sum_{i=1}^{m} \sum_{\alpha \in R_{+}} k_{w^{-1} \alpha}\left(w w^{-1} \alpha\right)_{i} \frac{f(w x)-f\left(w r_{w^{-1} \alpha} x\right)}{\left\langle x, w^{-1} \alpha\right\rangle} \\
& =w(\nabla f)(w x)+w \sum_{\alpha \in R_{+}} k_{\alpha} \alpha \frac{f(w x)-f\left(w r_{\alpha} x\right)}{\langle x, \alpha\rangle} \\
& =w \nabla_{k} f(w x),
\end{aligned}
$$

where we have changed the summation index to $w^{-1} \alpha$ in the second step.

To get an idea of the effect of the Dunkl operators, we look at effect of Dunkl operators on polynomials for some of the smaller root systems.

Example 6.7. Consider the root system $A_{1}$, with as only positive root $\alpha=\sqrt{2}$ and set $k_{\alpha}=k$. Then

$$
T_{1} x^{n}=n x^{n-1}+\sqrt{2} k \frac{x^{n}-(-1)^{n} x^{n}}{\sqrt{2} x}
$$

which gives

$$
T_{1} x^{n}=\left\{\begin{array}{cc}
(n+2 k) x^{n-1} & \text { for } n \text { is odd }, \\
n x^{n-1} & \text { for } n \text { is even. }
\end{array}\right.
$$

This shows that $T_{i} f=\partial_{i} f$ if the function is invariant under the reflection in $r_{\alpha}$.
Example 6.8. Consider the root system $B_{2}$, with weight function $k$, which has the positive roots $(\sqrt{2}, 0),(0, \sqrt{2}),(-1,1)$ and $(1,1)$. By the $G$-invariance of $k$ we must have that $k_{(\sqrt{2}, 0)}=$ $k_{(-\sqrt{2}, 0)}=l_{1}$ and $k_{(-1,1)}=k_{(1,1)}=l_{2}$.
Denote by $\rho(n), n \in \mathbb{N}$, the function which gives 0 if $n$ is even and 1 if $n$ is odd.
We use $(x, y)$ as basis on $\mathbb{R}^{2}$. From the previous example we see that

$$
s_{(\sqrt{2}, 0} x^{a} y^{b}=\sqrt{2} \rho(a) x^{a-1} y^{b}
$$

and

$$
s_{(0, \sqrt{2})} x^{a} y^{b}=\sqrt{2} \rho(b) x^{a} y^{b-1}
$$

For $a>b$, we have that

$$
s_{(-1,1)} x^{a} y^{b}=\frac{x^{a} y^{b}-x^{b} y^{a}}{x-y}=\sum_{i=0}^{a-b-1} \frac{x^{a-i} y^{b+i}-x^{a-1-i} y^{b+1+i}}{x-y}=\sum_{i=0}^{a-b-1} x^{a-i-1} y^{b+i}
$$

and in a similar way that

$$
s_{(1,1)} x^{a} y^{b}=\frac{x^{a} y^{b}-(-1)^{a+b} x^{b} y^{a}}{x+y}=\sum_{i=0}^{a-b-1} \frac{x^{a-i} y^{b+i}+x^{a-1-i} y^{b+1+i}}{(-1)^{i}(x+y)}=\sum_{i=0}^{a-b-1}(-1)^{i} x^{a-i-1} y^{b+i} .
$$

For $a<b$ we find these results with $x$ and $y$ interchanged and of course for $a=b$ both results are 0 . If we look at the the case where $a>b$, we see that

$$
T_{x}\left(x^{a} y^{b}\right)=\left(a-2 l_{1}\right) \rho(a) x^{a-1} y^{b}+-2 l_{2} \sum_{i=0}^{a-b-1} \rho(i) x^{a-i-1} y^{b+i}
$$

and

$$
T_{y}\left(x^{a} y^{b}\right)=\left(b-2 l_{1}\right) \rho(b) x^{a} y^{b-1}+2 l_{2} \sum_{i=0}^{a-b-1} \rho(i+1) x^{a-i-1} y^{b+i} .
$$

The results for $a \leq b$ are found in a similar way.

The main results of this chapter are Theorem 6.11 and Theorem 6.13. Theorem 6.11 states that

$$
T_{u} T_{v}=T_{v} T_{u}, \forall u, v \in \mathbb{R}^{m}
$$

Theorem 6.13 states that for each orthonormal basis of $\mathbb{R}^{m}$ the Dunkl Laplacian, which will be defined in Definition 6.12, can be written as

$$
\Delta_{k} f=\sum_{i=1}^{m} T_{i}^{2} f=\Delta f+2 k(\alpha) \sum_{\alpha \in R^{+}}\left(\frac{\langle\nabla f, \alpha\rangle}{\langle x, \alpha\rangle}-\frac{f(x)-f\left(r_{\alpha} x\right)}{\langle\alpha, x\rangle^{2}}\right),
$$

for $f \in C^{2}\left(\mathbb{R}^{m}\right)$.
To be able to prove these two theorems, we need some additional results.
The first result is

$$
\begin{equation*}
s_{\alpha} s_{\beta} f=\frac{f(x)}{\langle x, \alpha\rangle\langle x, \beta\rangle}-\frac{f\left(r_{\alpha} x\right)}{\langle x, \alpha\rangle\langle x, \beta\rangle}-\frac{f\left(r_{\beta} x\right)}{\langle x, \alpha\rangle\left\langle x, r_{\alpha} \beta\right\rangle}+\frac{f\left(r_{\alpha} r_{\beta} x\right)}{\langle x, \alpha\rangle\left\langle x, r_{\alpha} \beta\right\rangle},\left(f \in C\left(\mathbb{R}^{n}\right)\right) \tag{6.2}
\end{equation*}
$$

which follows from a straight-forward computation. [9, Eqn. 1.5]
The second result is the equation [9, Eqn. 1.6]

$$
\begin{align*}
\left\langle\nabla s_{\alpha} f, u\right\rangle-s_{\alpha}\langle\nabla f, u\rangle= & \frac{\langle\nabla f, u\rangle}{\langle x, \alpha\rangle}-\frac{\left\langle\nabla(f)\left(r_{\alpha} x\right), u\right\rangle}{\langle x, \alpha\rangle} \\
& -\frac{\langle\nabla f, u\rangle}{\langle x, \alpha\rangle}+\frac{\left\langle\nabla(f)\left(r_{\alpha} x\right), u\right\rangle}{\langle x, \alpha\rangle} \\
& -\langle u, \alpha\rangle \frac{f-f\left(r_{\alpha} x\right)}{\langle x, \alpha\rangle^{2}}+\frac{\left\langle\nabla f\left(r_{\alpha} x\right), \alpha\right\rangle\langle u, \alpha\rangle}{\langle x, \alpha\rangle} \\
= & \frac{\langle u, \alpha\rangle}{\langle x, \alpha\rangle}\left(s_{\alpha} f+\left\langle\nabla(f)\left(r_{\alpha} x\right), \alpha\right\rangle\right), \tag{6.3}
\end{align*}
$$

for $u \in \mathbb{R}^{m}, f \in C^{1}\left(\mathbb{R}^{m}\right)$.
Lemma 6.9. [9, Prop. 1.7] Let $B(x, y)$ be a bilinear form on $\mathbb{R}^{m}$, such that $B\left(r_{\alpha} x, r_{\alpha} y\right)=$ $B(y, x)$, when $\alpha \in \operatorname{span}(x, y)$. Let $w \in G$ be a plane rotation, which is a nontrivial product of two reflections. Then

$$
\text { (i) } \sum_{\substack{r_{\alpha}, r_{\beta} \in R_{+} \\ r_{\alpha} r_{\beta}=w}} k(\alpha) k(\beta) \frac{B(\alpha, \beta)}{\langle x, \alpha\rangle\langle x, \beta\rangle}=0 \text {, }
$$

when both sides are viewed as rational functions in $x$. Furthermore

$$
\text { (ii) } \sum_{\substack{r_{\alpha}, r_{\beta} \in R_{+} \\ r_{\alpha} r_{\beta}=w}} k(\alpha) k(\beta)\left(s_{\beta} s_{\alpha} B\right)(\alpha, \beta)=0,
$$

when both sides are viewed as functions $C^{2}\left(\mathbb{R}^{m}\right) \rightarrow C^{0}\left(\mathbb{R}^{m}\right)$.
Proof. Let $E$ be the plane of $w$, which means that $E$ is the plane orthogonal to the fixed point set of the action of $w$. If $r_{\alpha} r_{\beta}=w$, then $\alpha, \beta \in E$. Let $G_{1}$ be the subgroup of $G$ generated by $\left\{r_{\alpha} \mid \alpha \in E\right\}$. Let $m_{1}$ be the cardinality of the set of reflections in $G_{1}$. We also have that $r_{\alpha} w r_{\alpha}=w^{-1}$, if $r_{\alpha} \in G_{1}$.

Denote the sum in (i) by $t(x)$. We first want to show that $t(x)=G_{1}$-invariant.
For this, we fix a reflection $\sigma_{\gamma}$ in $G_{1}$ and define the functions $\epsilon(\alpha): R \rightarrow R$ and $\pi(\alpha): R \rightarrow$ $\{-1,1\}$ by $r_{\alpha} r_{\gamma} r_{\alpha}=r_{\epsilon(\alpha)}$ and $r_{\gamma} \alpha=\epsilon(\alpha) \pi(\alpha)$. So $\epsilon(\alpha)= \pm 1$ and $\epsilon(\pi(\alpha))=\epsilon(\alpha)$. Then

$$
\begin{align*}
t\left(r_{\gamma} x\right) & =\sum_{\substack{r_{\alpha}, r_{\beta} \in R_{+} \\
r_{\alpha} r_{\beta}=w}} k(\alpha) k(\beta) \frac{B(\alpha, \beta)}{\left\langle r_{\gamma} x, \alpha\right\rangle\left\langle r_{\gamma} x, \beta\right\rangle} \\
& =\sum_{\substack{r_{\alpha}, r_{\beta} \in R_{+} \\
r_{\pi}(\alpha) r^{\prime} r_{(\beta)}\\
}} k(\pi(\alpha)) k(\pi(\beta)) \frac{B(\pi(\alpha), \pi(\beta))}{\left\langle x, r_{\gamma} \pi(\alpha)\right\rangle\left\langle x, r_{\gamma} \pi(\beta)\right\rangle} \\
& =\sum_{\substack{r_{,}, r_{r} \in R_{+} \\
r_{\gamma} r_{\alpha} r_{\beta} r_{\gamma}=w}} k(\alpha) k(\beta) \frac{B\left(\epsilon(\alpha) r_{\gamma} \alpha, \epsilon(\beta) r_{\gamma} \beta\right)}{\left\langle r_{\gamma} x, \alpha\right\rangle\left\langle r_{\gamma} x, \beta\right\rangle \epsilon(\alpha) \epsilon(\beta)} \\
& =\sum_{\substack{\alpha_{,}, r_{r} \in R_{+} \\
r_{\alpha} r_{\beta}=r_{\gamma} w r_{\gamma}}} k(\alpha) k(\beta) \frac{B(\alpha, \beta)}{\left\langle r_{\gamma} x, \alpha\right\rangle\left\langle r_{\gamma} x, \beta\right\rangle} . \tag{6.4}
\end{align*}
$$

Because $G_{1}$ acts on a plane, we have that $\sigma_{\alpha} \sigma_{\beta}=\sigma_{\gamma} w \sigma_{\gamma}=w^{-1}$, if and only if $\sigma_{\beta} \sigma_{\alpha}=w$, so the last sum is equal to $t(x)$.
Next note that

$$
t(x) \prod_{\alpha \in E}\langle\alpha, x\rangle
$$

is a polynomial of degree $m_{1}-2$, which is alternating for $G_{1}$, hence it must be 0 and so $t(x)=0$.
To prove part (ii), we start with equation (6.2) and look at the terms of $f(x), f\left(r_{\gamma} x\right)$ and $f(w x)$. The coefficient of $f(x)$ is $t(x)$, so it is 0 . For a fixed $\gamma \in E$, the coefficient of $f\left(r_{\gamma} x\right)$ is

$$
\frac{k_{\alpha} k_{\gamma} B(\alpha, \gamma)}{\langle x, \alpha\rangle\langle x, \gamma\rangle}+\frac{k_{\beta} k_{\gamma} B(\gamma, \beta)}{\langle x, \gamma\rangle\left\langle x, r_{\gamma} \beta\right\rangle},
$$

where $r_{\gamma} r_{\beta}=w=r_{\beta} r_{\gamma}$. By defining the functions $\epsilon$ and $\pi$ as before, we have $\beta=\pi(\alpha)$ and the second term can be rewritten as

$$
\frac{k_{\pi(\alpha)} k_{\gamma} B\left(\gamma, \epsilon(\alpha) r_{\gamma} \alpha\right)}{\epsilon(\alpha)\langle x, \gamma\rangle\langle x, \alpha\rangle}=\frac{k_{\alpha} k_{\gamma} B\left(\alpha, r_{\gamma} \gamma\right)}{\langle x, \gamma\rangle\langle x, \alpha\rangle} .
$$

So the coefficient is zero since $r_{\gamma} \gamma=-\gamma$.
The coefficient of $f(x w)$ is

$$
\sum_{r_{\alpha} r_{\beta}=w} \frac{k_{\alpha} k_{\beta} B(\alpha, \beta)}{\langle x, \alpha\rangle\langle x, \beta\rangle} .
$$

For a fixed $\alpha \in E$, there is a unique $\beta \in E$, such that $r_{\alpha} r_{\beta}=w$. Let $r_{\delta}=r_{\alpha} r_{\beta} r_{\alpha}$, which means that $\delta=\epsilon(\beta) r_{\alpha}(\beta)$. Then the $(\alpha, \beta)$-term equals

$$
\frac{k_{\alpha} k_{\beta} B\left(r_{\alpha} \beta, r_{\alpha} \alpha\right)}{\langle x, \alpha\rangle\langle x, \epsilon(\beta) \delta\rangle}=\frac{-k_{\alpha} k_{\delta} B(\delta, \alpha)}{\langle x, \delta\rangle\langle x, \alpha\rangle} .
$$

Since $r_{\delta} r_{\alpha}=w$, the sum equals $-t(x)$, so all the coefficients are zero and we have proven both parts of the lemma.

Corollary 6.10. [9, Cor. 1.8] Let $B(x, y)$ be a bilinear form on $\mathbb{R}^{m}$, such that $B\left(r_{\alpha} x, r_{\alpha} y\right)=$ $B(y, x)$, when $\alpha \in \operatorname{span}(x, y)$. Then

$$
\sum_{\alpha, \beta \in R_{+}} k(\alpha) k(\beta)\left(s_{\beta} s_{\alpha} B\right)(\alpha, \beta)=0
$$

Proof. The terms with $\alpha=\beta$ are 0 , since $s_{\alpha}^{2}=0$. The other terms can be grouped by the value of $r_{\alpha} r_{\beta}$, which is a plane rotation. Each of these groups of terms, sums to zero, because of part (ii) of Lemma 6.9.

Finally we can prove the main results of this chapter.
Theorem 6.11. [9, Thm. 1.9] Let $u, v$ be two vectors in $\mathbb{R}^{m}$. Then $T_{u} T_{v}=T_{v} T_{u}$, when viewed as operators on $C^{2}\left(\mathbb{R}^{m}\right)$.

Proof. Expand $\left(T_{u} T_{v}-T_{v} T_{u}\right) f=E_{1}+E_{2}+E_{3}$, with

$$
\begin{gathered}
E_{1}=\langle\nabla\langle u, \nabla f(x)\rangle, v\rangle-\langle\nabla\langle u, \nabla f(x)\rangle, v\rangle=0, \\
E_{2}=\sum_{\alpha \in R_{+}} k_{\alpha}\langle v, \alpha\rangle\left(\left\langle\nabla s_{\alpha} f(x), u\right\rangle-s_{\alpha}\langle\nabla f(x), u\rangle\right)-\sum_{\alpha \in R_{+}} k_{\alpha}\langle u, \alpha\rangle\left(\left\langle\nabla s_{\alpha} f(x), v\right\rangle-s_{\alpha}\langle\nabla f(x), v\rangle\right)
\end{gathered}
$$

and

$$
E_{3}=\sum_{\alpha, \beta \in R_{+}} k_{\alpha} k_{\beta} s_{\beta} s_{\alpha} B(\alpha, \beta) f(x),
$$

with

$$
B(x, y)=\langle u, x\rangle\langle v, y\rangle-\langle u, y\rangle\langle v, x\rangle .
$$

The operators in $E_{1}$ are the usual partial derivatives which commute. Since $B(x, y)$ satisfies the hypothesis of Corollary 6.10, $E_{3}$ is zero.
By using Equation (6.3), we see that

$$
E_{2}=\sum_{\alpha} k_{\alpha}(\langle v, \alpha\rangle\langle u, \alpha\rangle-\langle u, \alpha\rangle\langle v, \alpha\rangle) \times\left(2\left\langle\alpha, \nabla f\left(r_{\alpha} x\right)-s_{\alpha} f(x)\right) /\langle x, \alpha\rangle=0,\right.
$$

so $T_{u}$ and $T_{v}$ commute.
Definition 6.12. [9, Thm. 1.10] For a given root system $R_{+}$with weight function $k$, and an orthonormal basis $e_{i}, 1 \leq i \leq m$ of $\mathbb{R}^{m}$, we define the Dunkl Laplacian $\Delta_{k}: C^{2}\left(\mathbb{R}^{m}\right) \rightarrow$ $C^{0}\left(\mathbb{R}^{m}\right)$ by

$$
\Delta_{k} f=\sum_{i=1}^{m} T_{i}^{2} f
$$

For $k=0$, the operator $\Delta_{k}$ coincides with the ordinary Laplacian in $m$ variables.
Theorem 6.13. [9, Thm. 1.10] For any orthonormal basis of $\mathbb{R}^{m}$ the Dunkl Laplacian is given by

$$
\Delta_{k} f=\Delta f+2 \sum_{\alpha} k_{\alpha}\left(\frac{\langle\nabla f, \alpha\rangle}{\langle x, \alpha\rangle}-\frac{f(x)-f\left(r_{\alpha} x\right)}{\langle\alpha, x\rangle^{2}}\right) .
$$

In particular $\Delta_{k}$ is independent of this basis.

Proof. By using Definition 6.3 we find

$$
\begin{aligned}
T_{u}^{2}= & \langle\nabla\langle\nabla f(x), u\rangle, u\rangle+2 \sum_{\alpha} k_{\alpha}\langle u, \alpha\rangle\langle u, \nabla f(x)\rangle /\langle x, \alpha\rangle \\
& -\sum_{\alpha} k_{\alpha}\langle u, \alpha\rangle^{2}\left(f(x)-f\left(r_{\alpha}(x)\right)\right) /\langle x, \alpha\rangle^{2} \\
& -\sum_{\alpha} k_{\alpha}\langle u, \alpha\rangle\left(2\left\langle u, \nabla f\left(r_{\alpha}(x)\right)\right\rangle-\langle u, \alpha\rangle\left\langle\alpha, \nabla f\left(r_{\alpha} x\right)\right\rangle\right) /\langle x, \alpha\rangle \\
& +\sum_{\alpha, \beta} k_{\alpha} k_{\beta}\langle u, \alpha\rangle\langle u, \beta\rangle\left(s_{\alpha} s_{\beta} f\right)(x) .
\end{aligned}
$$

We use Definition 6.12 and Parsevals identity $\sum_{i=1}^{m}\left\langle e_{i}, u\right\rangle\left\langle e_{i}, v\right\rangle=\langle u, v\rangle$, where $e_{i}, 1 \leq i \leq m$ is an orthonormal basis on $\mathbb{R}^{m}$, to find

$$
\begin{aligned}
\Delta_{k}(f)=\sum_{i=1}^{m} T_{e_{i}}^{2}= & \Delta f(x)+2 \sum_{\alpha} k_{\alpha}\langle\alpha, \nabla f(x)\rangle /\langle x, \alpha\rangle \\
& -\sum_{\alpha} k_{\alpha} 2\left(f(x)-f\left(r_{\alpha}(x)\right)\right) /\langle x, \alpha\rangle^{2} \\
& -\sum_{\alpha} k_{\alpha}\left(2\left\langle\alpha, \nabla f\left(r_{\alpha}(x)\right)\right\rangle-2\left\langle\alpha, \nabla f\left(r_{\alpha} x\right)\right\rangle\right) /\langle x, \alpha\rangle \\
& +\sum_{\alpha, \beta} k_{\alpha} k_{\beta}\langle\alpha, \beta\rangle\left(s_{\alpha} s_{\beta} f\right)(x) \\
= & \Delta f+2 \sum_{\alpha} k_{\alpha}\left(\frac{\langle\nabla f, \alpha\rangle}{\langle x, \alpha\rangle}-\frac{f(x)-f\left(r_{\alpha} x\right)}{\langle\alpha, x\rangle^{2}}\right)
\end{aligned}
$$

because the last sum is zero by applying Corollary 6.10 to the form $B(x, y)=\langle x, y\rangle$.

### 6.1 The Dunkl harmonic Fischer decomposition

Since we have defined the Dunkl Laplacian $\Delta_{k}$, we can generalize the harmonic Fischer decomposition to the setting of Dunkl operators. First define $\gamma=\sum_{\alpha \in r_{+}} k_{\alpha}$ and define the Dunkl dimension by $m_{k}=m+2 \gamma$. Also recall that $E=\sum_{i=1}^{m} x_{i} \partial / \partial x_{i}$ is the Euler operator as given in Chapter 2 .
Lemma 6.14. [18, Thm. 3.3] The linear span of the operators $|x|^{2}, \Delta_{k}$ and $E$ in $\operatorname{End}\left(C^{\infty}\left(\mathbb{R}^{m}\right)\right)$, equipped with the commutator bracket is a Lie algebra isomorphic to $\mathfrak{s l}_{2}$.
Proof. For $\Delta_{k}$ and $|x|^{2}$ we find

$$
\begin{aligned}
{\left[\Delta_{k},|x|^{2}\right] f } & =\Delta_{k}\left(|x|^{2} f\right)-|x|^{2} \Delta_{k} f \\
& =\Delta\left(|x|^{2} f\right)-|x|^{2} \Delta(f)+2 \sum_{\alpha} k_{\alpha} \frac{\left.\left.\left\langle\nabla\left(|x|^{2} f\right)-\right| x\right|^{2} \nabla(f), \alpha\right\rangle}{\langle x, \alpha\rangle} \\
& -\sum_{\alpha} k_{\alpha} \frac{|x|^{2} f(x)-r_{\alpha}(|x|)^{2} f\left(r_{\alpha} x\right)-|x|^{2} f(x)+|x|^{2} f\left(r_{\alpha}(x)\right)}{\langle x, \alpha\rangle^{2}} \\
& =4(E+m / 2)+2 \sum_{\alpha} k_{\alpha} \frac{\langle x f, \alpha\rangle}{\langle x, \alpha\rangle}-0 \\
& =4(E+m / 2+\gamma)=4\left(E+m_{k} / 2\right),
\end{aligned}
$$

where we have used that $r_{\alpha}|x|^{2}=|x|^{2}$ and that $\left[\Delta,|x|^{2}\right]=4(E+m / 2)$.
For $E+m_{k} / 2$ and $|x|^{2}$ we find

$$
\begin{aligned}
{\left[E+m_{k} / 2,|x|^{2}\right] f } & =\left(E|x|^{2}-|x|^{2} E\right) f \\
& =\sum_{i=1}^{m}\left(|x|^{2} x_{i} \partial_{i}-|x|^{2} x_{i} \partial_{i}+\partial_{i}\left(|x|^{2}\right) x_{i}\right) f \\
& =\sum_{i=1}^{m} 2 x_{i}^{2} f \\
& =2|x|^{2} f
\end{aligned}
$$

To compute the commutator of $E+m_{k} / 2$ and $\Delta_{k}$, we need the following results.

$$
\begin{aligned}
{\left[E,\langle x, \alpha\rangle^{-1} \nabla_{\alpha}\right] f } & =\sum_{i=1}^{m} x_{i} \partial_{i} \frac{\langle\nabla f, \alpha\rangle}{\langle x, \alpha\rangle}-\sum_{i=1}^{m} \frac{\left\langle\nabla\left(x_{i} \partial_{i} f\right), \alpha\right\rangle}{\langle x, \alpha\rangle} \\
& =\sum_{i, j=1}^{m} \frac{\alpha_{j}\langle x, \alpha\rangle x_{i} \partial_{i} \partial_{j} f-\alpha_{i} \alpha_{j} x_{i} \partial_{j} f}{\langle x, \alpha\rangle^{2}}-\sum_{i, j=1}^{m} \frac{\alpha_{j} \partial_{j}\left(x_{i}\right) \partial_{i} f+\alpha_{j} x_{i} \partial_{i} \partial_{j} f}{\langle x, \alpha\rangle} \\
& =\frac{E\langle\nabla f, \alpha\rangle-\langle\nabla f, \alpha\rangle-\langle\nabla f, \alpha\rangle-E\langle\nabla f, \alpha\rangle}{\langle x, \alpha\rangle} \\
& =-2 \frac{\langle\nabla f, \alpha\rangle}{\langle x, \alpha\rangle}
\end{aligned}
$$

where we have used the quotient rule for derivations.
Also note that

$$
\begin{aligned}
{\left[E,\langle x, \alpha\rangle^{-1} s_{\alpha}\right] f } & =E \frac{f(x)-f\left(r_{\alpha} x\right)}{\langle x, \alpha\rangle^{2}}-\frac{(E f)(x)-(E f)\left(r_{\alpha} x\right)}{\langle x, \alpha\rangle^{2}} \\
& =\frac{\langle x, \alpha\rangle^{2}\left((E f)(x)-(E f)\left(r_{\alpha} x\right)\right)-2\langle x, \alpha\rangle^{2}\left(f(x)-f\left(r_{\alpha} x\right)\right)}{\langle x, \alpha\rangle^{4}} \\
& -\frac{(E f)(x)-(E f)\left(r_{\alpha} x\right)}{\langle x, \alpha\rangle^{2}} \\
& =-2 \frac{f(x)-f\left(r_{\alpha} x\right)}{\langle x, \alpha\rangle^{2}} \\
& =-2\langle x, \alpha\rangle^{-1} s_{\alpha}
\end{aligned}
$$

where we have used the quotient rule again.
By the previous equations and the result that $[E, \Delta]=-2 \Delta$, we find that

$$
\begin{aligned}
{\left[E, \Delta_{k}\right] } & =[E, \Delta]+\sum_{\alpha \in R+} k_{\alpha}\left(\left[E,\langle x, \alpha\rangle^{-1} \nabla_{\alpha}\right]-\left[E,\langle x, \alpha\rangle^{-1} s_{\alpha}\right]\right) \\
& =-2 \Delta-2 \sum_{\alpha \in R+} k_{\alpha}\langle x, \alpha\rangle^{-1}\left(\nabla_{\alpha}-s_{\alpha}\right) \\
& =-2 \Delta_{k}
\end{aligned}
$$

Lemma 6.15. [8, Lemma 1.9] Let $l \neq 0 \in \mathbb{N}$. The operator $\Delta_{k}|x|^{2}$ acts as a nonzero scalar on the spaces $|x|{ }^{2 l} H_{k, n}$. In particular, its action is invertible.

Proof. This proof is based on Leibniz rule

$$
\left[A^{l}, E\right]=\sum_{j=1}^{l-1} A^{j-1}[A, E] A^{l-j}
$$

Let $h_{n} \in H_{k, n}$. When we apply the Leibniz rule with $A=|x|^{2}$ we get

$$
\begin{align*}
\Delta|x|^{2 l} h_{n} & =|x|^{2 l} \Delta h_{n}+\sum_{i=0}^{l-1}|x|^{2(l-1-i)}\left[|x|^{2}, E\right]|x|^{2 i} h_{n} \\
& =|x|^{2 l} \Delta h_{n}+\sum_{i=0}^{l-1}|x|^{2(l-1-i)} 4\left(E+m_{k} / 2\right)|x|^{2(i)} h_{n} \\
& =\sum_{i=0}^{l-1}|x|^{2(l-i-1)} 4\left(2 i+n+m_{k} / 2\right)|x|^{2(i)} h_{n} \\
& =\sum_{i=0}^{l-1} 4\left(2 i+n+m_{k} / 2\right)|x|^{2(l-1)} h_{n} \\
& =4(l)\left(n+m_{k} / 2+l-1\right)|x|^{2(l-1)} h_{n} \\
& :=c_{k, n l}|x|^{2(l-1)} h_{n}, \tag{6.5}
\end{align*}
$$

because $k_{\alpha}$ is positive the constants $c_{k, n l}$ are nonzero.
Theorem 6.16. [8, Thm. 1.7] Let $R$ be a root system in $\mathbb{R}^{m}$ with weight function $k$. Assume that $k$ is positive. Denote by $H_{k, n}$ the spaces of homogeneous $k$-harmonic functions of degree $n$. Then $P_{n}\left(\mathbb{R}^{m}\right)$ can be decomposed as

$$
P_{n}=\oplus_{i \leq n / 2}|x|^{2 i} H_{k, n-2 i} .
$$

We call this decomposition the Dunkl harmonic Fischer decomposition.
Proof. We can modify the proof of Theorem 2.1. When we use this proof with the constants $c_{k, n l}$ instead of the constants $c_{n l}$, this proof leads to the Dunkl harmonic Fischer decomposition.

Remark 6.17. We can construct a basis of $H_{k, n}$ by using the Dunkl harmonic Fischer decomposition. This is basically done by applying the steps given in Section 3 .

### 6.2 The decomposition of $L^{2}\left(S, h^{2} d \omega\right)$

In this section we will restrict the Dunkl harmonic Fischer decomposition from Theorem 6.16 to the sphere $S$, and show that this gives an orthogonal decomposition of $L^{2}\left(S, h^{2} d \omega\right)$. For this we use results from Dunkl, which were given in [8, Ch.1]. In particular we need a generalization of Green's theorem, to prove an orthogonality result on the $k$-harmonic polynomials.

First we need some additional definitions.

Definition 6.18. Define as before the function $h: \mathbb{R}^{m} \rightarrow \mathbb{R}$ by

$$
h(x)=\prod_{\alpha \in R_{+}}|\langle\alpha, x\rangle|^{k_{\alpha}} .
$$

Define $\gamma=\operatorname{deg}(h)$ and define the Dunkl dimension $m_{k}$ by $m_{k}=m+2 \operatorname{deg} h$.
The function $h(x)$ is $G$-invariant, since each element of $G$ interchanges the roots and $k_{\alpha}=k_{\beta}$, if $\alpha$ and $\beta$ are conjugate. As we shall see, the Dunkl dimension occurs instead of the dimension $m$ in generalizations of the harmonic analysis. This can already be seen by comparing the constants in equations (2.1) and (6.5).

Definition 6.19. Define by $d \omega$ the normalized rotation invariant surface measure on the sphere $S=\left\{x \in \mathbb{R}^{m}:|x|=1\right\}$ and by $d x$ the Lebesgue measure on $\mathbb{R}^{m}$. We shall use the measure $h^{2} d \omega$ on the unit sphere, the measure $h^{2} d x$ on $\mathbb{R}^{m}$ and the Gaussian measure

$$
h^{2} d \mu=h^{2}(x)(2 \pi)^{-m / 2} e^{-|x|^{2} / 2} d x
$$

on $\mathbb{R}^{m}$.
We define the normalization constants $c_{m}$ and $c_{m}^{\prime}$ by $c_{m}=\left(\int_{\mathbb{R}^{m}} h^{2} d \mu\right)^{-1}$ and $c_{m}^{\prime}=\left(\int_{S^{m-1}} h^{2} d \omega\right)^{-1}$.
If $f$ is a continuous function of polynomial growth on $\mathbb{R}^{n}$, then by using polar coordinates,

$$
\int_{\mathbb{R}^{n}} f d \mu=\left(2^{1-m / 2} / \Gamma\left(\frac{m}{2}\right)\right) \int_{0}^{\infty} \int_{S} r^{m-1} e^{-r^{2} / 2} f(r x) d \omega(x) d r .
$$

If $f$ is positively homogeneous of degree $2 k$, then

$$
\int_{\mathbb{R}^{n}} f h^{2} d \mu=2^{\gamma+k}\left(\Gamma\left(\frac{m}{2}+k+\gamma\right) / \Gamma\left(\frac{m}{2}\right)\right) \int_{S} f h^{2} d \omega
$$

By combining these results with the normalization constants we find ([11, p. 1215])

$$
\begin{equation*}
c_{m}^{\prime}=2^{\gamma}\left(\Gamma\left(\frac{m}{2}+\gamma\right) / \Gamma\left(\frac{m}{2}\right)\right) c_{m} \tag{6.6}
\end{equation*}
$$

We can split the Laplacian $\Delta_{k}: C^{2}\left(\mathbb{R}^{m}\right) \rightarrow C^{0}\left(\mathbb{R}^{m}\right)$, which was given in 6.12, as $\Delta_{k}=L_{k}-D_{k}$, were $L_{k}$ is the differential part

$$
L_{k} f=\Delta f+\sum_{\alpha \in R_{+}} 2 k_{\alpha} \frac{\langle\alpha, \nabla f\rangle}{\langle x, \alpha\rangle},
$$

and $D_{k}$ is the difference part

$$
D_{k}=\sum_{\alpha \in R_{+}} 2 k_{\alpha} \frac{f(x)-f\left(r_{\alpha}(x)\right)}{\langle x, \alpha\rangle^{2}} .
$$

Note that $L_{k}$ can be written as $L_{k}=(\Delta(f h)-f \Delta h) / h$ (see [8, Propositions 1.1,1.3]).
Lemma 6.20. [8, Prop. 1.2] The operator $\left.D_{k}\right|_{S}$ is symmetric on $L^{2}\left(S, h^{2} d \omega\right)$.

Proof. For each $k_{\alpha}>1,\left(\alpha \in R_{+}\right)$and for each $f, g \in L^{2}\left(S, h^{2} d \omega\right)$, the function $f g h^{2}\langle x, \alpha\rangle^{-2}$ is integrable. Also note that the reflection $r_{\alpha}$ sends $\alpha$ to $-\alpha$ and interchanges the other positive roots. This gives

$$
\begin{aligned}
\int_{S} D_{k}(f) g h^{2} d \omega & =\sum_{\alpha \in R_{+}} 2 k_{\alpha}\left(\int_{S} \frac{f(x) g(x)}{\langle x, \alpha\rangle^{2}} h(x)^{2} d \omega-\int_{S} \frac{f\left(r_{\alpha} x\right) g(x)}{\langle x, \alpha\rangle^{2}} h(x)^{2} d \omega\right) \\
& =\sum_{\alpha \in R_{+}} 2 k_{\alpha}\left(\int_{S} \frac{f(x) g(x)^{2}}{\langle x, \alpha\rangle} h(x)^{2} d \omega-\int_{S} \frac{f(x) g\left(r_{\alpha} x\right)^{2}}{\langle x, \alpha\rangle} h(x)^{2} d \omega\right) \\
& =\int_{S} f D_{k}(g) h^{2} d \omega,
\end{aligned}
$$

where we have changed the integration variable to $r_{\alpha} x$ in the second sum. This is valid, since $h$, the measure $d \omega$ and the space $S$ are $G$-invariant.

Remark 6.21. Let $B=\left\{x \in \mathbb{R}^{m} \mid x<1\right\}$ be the open ball, with closure $\bar{B}=B \cap S$. Note that $\bar{B}$ is invariant under reflections and the function $\operatorname{fgh}^{2}\langle x, \alpha\rangle^{-2}$ is integrable on $\bar{B}$, for $f, g \in C^{2}(\bar{B})$. The operator $\left.D_{k}\right|_{\bar{B}}$ is symmetric on $L^{2}\left(\bar{B}, h^{2} d x\right)$, by an argument similar to proof of Lemma 6.20.

Theorem 6.22. [8, Prop. 1.4] Let $B=\left\{x \in \mathbb{R}^{m}:|x|<1\right\}$ be the open ball, with measure $h^{2} d x$, and let $f, g$ be $C^{2}$ functions on its closure $\bar{B}=B \cup S$. Let $S$ have the surface measure $h^{2} d \omega$. Denote by $\eta$ the outward normal vector on $S$ and denote by $c$ the normalization constant $\left(\Gamma(m / 2)(2 \pi)^{m / 2}\right)^{-1}$. Then

$$
c \int_{S} \frac{\partial f}{\partial \eta} g h^{2} d \omega=\int_{B}\left(g L_{k}(f)-\langle\nabla f, \nabla g\rangle\right) h^{2} d x .
$$

Proof. Green's identity gives

$$
c \int_{S} \frac{\partial f_{1}}{\partial \eta} f_{2} d \omega=\int_{B} f_{2} \Delta f_{1}+\left\langle\nabla f_{1}, \nabla f_{2}\right\rangle d x
$$

for $f_{1}, f_{2} \in C^{2}(\bar{B})$. If we apply this to $f_{1}=f h$ and $f_{2}=g h$, we find

$$
c \int_{S}\left(\frac{\partial f}{\partial \eta} g h^{2}+\frac{\partial h}{\partial \eta} f g h\right) d \omega=\int_{B}(g h \Delta(f h)+\langle\nabla(f h), \nabla(g h)\rangle) d x .
$$

Applying Green's identity to $f_{1}=h$ and $f_{2}=f g h$ gives

$$
c \int_{S} \frac{\partial h}{\partial \eta} f g h d \omega=\int_{B}(f g h \Delta(h)+\langle\nabla(f g h), \nabla(h)\rangle) d x
$$

Now we can substract these two equations from each other and find, using the product rule $\langle\nabla(f g), \nabla(h)\rangle=f\langle\nabla g, \nabla h\rangle+g\langle\nabla f, \nabla h\rangle$, that

$$
\begin{aligned}
c \int_{S} \frac{\partial f}{\partial \eta} g h^{2} d \omega & =\int_{B}(g h \Delta(f h)-g h(f \Delta h)+\langle\nabla(f h), \nabla(g h)\rangle-\langle\nabla(f g h), \nabla h\rangle) d x \\
& =\int_{B}\left(g L_{k}(f)+h^{2}\langle\nabla f, \nabla g\rangle\right) d x
\end{aligned}
$$

Lemma 6.23. [8, Theorem 1.6] Let $f$ and $g$ be homogeneous $k$-harmonic polynomials of different degree, then

$$
\int_{S} f g h^{2} d \omega=0
$$

Proof. Using polar coordinates we have for $f \in P_{n}\left(\mathbb{R}^{m}\right)$

$$
C_{n} \int_{B} f(x) d x=\int_{0}^{1} r^{n-1} d r \int_{S} f(x) d \omega(x)=1 / \operatorname{deg}(f) \int_{S} f(x) d \omega(x)
$$

so

$$
\begin{aligned}
(\operatorname{deg}(f)-\operatorname{deg}(g)) \int_{S} f g h^{2} d \omega & =\int_{S}(\partial f / \partial \eta) g h^{2} d \omega-\int_{S} f(\partial g / \partial \eta) h^{2} d \omega \\
& =\int_{\bar{B}}\left(g L_{k} f-f L_{k} g\right) h^{2} d x \\
& =\int_{\bar{B}}\left(g\left(L_{k}-D_{k}\right) f-f\left(L_{k}-D_{k}\right) g\right) h^{2} d x \\
& =0
\end{aligned}
$$

where we have used that $D_{k}$ is symmetric on $L^{2}\left(\bar{B}, h^{2} d x\right)$ by Remark 6.21 . So for $\operatorname{deg}(f) \neq$ $\operatorname{deg}(g)$ we see that $\int f g h^{2} d \omega=0$.

Corollary 6.24. [8, p. 39] Restriction of the Dunkl harmonic Fischer decomposition, which was defined in Theorem 6.16, leads to the decomposition

$$
L^{2}\left(S^{m}, h^{2} d \omega\right)=\left.\widehat{\bigoplus}_{n=0}^{\infty} H_{k, n}\right|_{S}
$$

Proof. In Theorem 6.22 is shown that $\left.\left.H_{k, n}\right|_{S} \perp H_{k, l}\right|_{S}$ for $n \neq l$. Since $|x|^{2 n}=1$ on the unit sphere, we have that

$$
\left.P\right|_{S}=\left.\sum_{n=0}^{\infty} \sum_{i=0}^{\lfloor n / 2\rfloor}\left(|x|^{2 i} H_{k, n-2 i}\right)\right|_{S}=\left.\left.P\left(|x|^{2}\right)\right|_{S} \bigotimes \sum_{n=0}^{\infty} H_{k, n}\right|_{S}=\left.\sum_{n=0}^{\infty} H_{k, n}\right|_{S}
$$

where we have denoted the space of all polynomials in $|x|^{2}$ by $P\left(|x|^{2}\right)$. Since $\left.\left.H_{k, n}\right|_{S} \perp H_{k, l}\right|_{S}$ for $n \neq l$, the sum $\sum_{n=0}^{\infty} H_{k, n}$ is orthogonal. In particular, it is direct.
By Stone-Weierstrass the space $\left.P\right|_{S}$ is dense $C(S)$, so it is also dense in $L^{2}\left(S^{m}, h^{2} d \omega\right)$. This gives the decomposition

$$
L^{2}\left(S^{m}, h^{2} d \omega\right)=\widehat{\oplus}_{n=0}^{\infty} H_{k, n} \mid S
$$

Theorem 6.25. [8, Thm. 1.6] Let $p$ be an element of $P_{n}$. Then

$$
\int p q h^{2} d \omega=0, \text { for all } q \in \sum_{i=1}^{n-1} P_{i}
$$

if and only if $\Delta_{k} p=0$.

Proof. Let $p$ be an element of $P_{n}$ and write $p$ as

$$
p=\sum_{j} p_{n-2 j}|x|^{2 j},
$$

with $p_{n-2 j} \in H_{n-2 j}$. Now suppose that $p$ is not k -harmonic, so there is some $j \neq 0$ such that $p_{n-2 j} \neq 0$. Then for $q=p_{n-2 j}$ the integral $\int p q h^{2} d \omega=\int p_{n-2 j}^{2} h^{2} d \omega \neq 0$, because of positivity of the inner product. So if $p$ is not $k$-harmonic there is some polynomial of lower degree such that $\int_{S} p q h^{2} d w \neq 0$. If $p$ is harmonic, choose $q \in P_{i}, i<n$ arbitrary. By restricting $q$ to $S$, we see that $q \in \sum_{l=0}^{i} H_{l}$, so by Theorem 6.22, we have that

$$
\int p q h^{2} d w=0 .
$$

Theorem 6.26. [8, p. 38] Let $j, l, n \in \mathbb{N}$. The operator $|x|^{2 l} \Delta_{k}^{l}$ acts on the space $|x|^{2 j} H_{k, n}$ by the scalar

$$
\lambda_{n j}^{l}=4^{l}(-j)_{l}\left(-n-m_{k} / 2-j+1\right)_{l} .
$$

Here $(j)_{l}$ is the Pochhammer symbol, given by $(j)_{l}=j(j+1)(j+2) \ldots(j+l-1)$.
Proof. By (6.5), we have for $h_{n} \in H_{k, n}$ that

$$
\begin{equation*}
\Delta_{k}|x|^{2 j} h_{n}(x)=4(j)\left(n+m_{k} / 2+j-1\right)|x|^{2(j-1)} h_{n}(x) . \tag{6.7}
\end{equation*}
$$

By repeating this process we find

$$
\begin{aligned}
|x|^{2 l} \Delta_{k}^{l}|x|^{2 j} h_{n}(x) & =\prod_{i=1}^{l} 4(k-i+1)\left(n+m_{k} / 2+k-i\right)|x|^{2 j} h_{n}(x) \\
& =4^{l}(-j)_{l}\left(-n-m_{k}-k+1\right)_{l}|x|^{2 j} h_{n}(x)
\end{aligned}
$$

Note that $|x|^{2 l} \Delta_{k}^{l} r^{2 j} h_{n}=0$ for $l>j$, because $(-j)_{l}=0$ for $l>j$.
We can use the constants above, to give an the Dunkl harmonic Fischer decomposition in an explicit way. This was already done in slightly different ways in [8, Thm. 1.11] and [20, Cor. 4.1].

Definition 6.27. Define the operators $Q_{n, l}: P \rightarrow P$ by

$$
Q_{n, l}=1-\frac{|x|^{2} \Delta_{k}}{\lambda_{n-2 l, l}^{1}},
$$

The operators $Q_{n, l}$ have the same eigenspace decomposition as $|x|^{2} \Delta_{k}$ but they have different eigenvalues. In particular for $h_{2 n-l} \in H_{k, 2 n-l}$ we find $Q_{n, l} \left\lvert\, x{ }^{2 l} h_{2 n-l}=1-\frac{4 l\left(n-l+m_{k} / 2-1\right.}{4 l\left(n-l+m_{k} / 2-1\right)}=0\right.$.
Theorem 6.28. Let $f \in P_{n}$. Let $f=\sum_{j=1}^{\lfloor n / 2\rfloor}|x|^{2 j} f_{j}$ with $f_{j} \in H_{n-2 j}$ according to the Fischer decomposition. Then $f_{j}$ is given by

$$
|x|^{2 j} f_{j}=\left(\prod_{l=j+1}^{\lfloor n / 2\rfloor} Q_{n, l}\right) \frac{|x|^{2 j} \Delta_{k}^{j}}{\lambda} f,
$$

with

$$
\lambda=\prod_{l=j+1}^{\lfloor n / 2\rfloor} \frac{\lambda_{n-2 l, l}^{1}-\lambda_{n-2 j, j}^{1}}{\lambda_{n-2 l, l}^{1}} \lambda_{n-2 j, j}^{i} .
$$

Proof. Note that $x^{2 j} \Delta_{j} x^{2 i} f_{i}=0$ for $i<j$ and $x^{2 j} \Delta_{j} x^{2 i} f_{i}=\lambda_{n-2 i, i}^{i} x^{2 i} f_{i}$ by Theorem 6.26. By Theorem 6.27, we have $Q_{n, i}|x|^{2 i} f_{i}=0$ and $Q_{n, i}|x|^{2 l} f_{l}=c|x|{ }^{2 l} f_{l}$, with c some real constant depending on $l, i$ and $n$. By putting these results together, we find that

$$
\left(\prod_{l=j+1}^{\lfloor n / 2\rfloor} Q_{n, l}\right)|x|^{2 j} \Delta_{k}^{j} f=\lambda|x|^{2 j} f_{j},
$$

where $\lambda$ is some real constant which needs to be computed. A simple computation shows that

$$
\left(\prod_{l=j+1}^{\lfloor n / 2\rfloor} Q_{n, l}\right)|x|^{2 j} \Delta_{k}^{j}|x|^{2 j} f_{j}=\prod_{l=j+1}^{\lfloor n / 2\rfloor} \frac{\lambda_{n-2 l, l}^{1}-\lambda_{n-2 j, j}^{1}}{\lambda_{n-2 l, l}^{1}} \lambda_{n-2 j, j}^{i} f_{j}
$$

So

$$
\lambda=\prod_{l=j+1}^{\lfloor n / 2\rfloor} \frac{\lambda_{n-2 l, l}^{1}-\lambda_{n-2 j, j}^{1}}{\lambda_{n-2 l, l}^{1}} \lambda_{n-2 j, j}^{i} .
$$

These constants can also be found in a slightly different way. We can view the polynomials $x^{2 l} \Delta_{k}^{l} f$ as the solution of a system of equations in the unknowns $|x|^{2 j} f_{j}$. We can write this in matrixform as

$$
\Gamma_{j l}|x|^{2 j} f_{j}=x^{2 l} \Delta_{k}^{l} f,
$$

for $0 \leq j, l \leq\lfloor n / 2\rfloor$ with $\Gamma_{j l}=\lambda_{n-2 j, j}^{l}$. Since the matrix $\Gamma$ is upper triangular with non-zero diagonal entries, we can solve it by Gaussian elimination.

Lemma 6.29. Let $x, y \in \mathbb{R}^{n}$ and let $V$ be an uppertriangular $n \times n$ matrix, with nonzero diagonal entries. Then we can solve the system $V \cdot x=y$ by Gaussian elimination in particular we have

$$
v_{i i} x_{i}=y_{i}+\sum_{a=i+1}^{n} \sum_{b=a}^{n}-\frac{v_{i b}}{v_{b b}}\left(\prod_{c=b+1}^{n}-\frac{v_{c-1, c}}{v_{c, c}}\right) y_{b} .
$$

Proof. To find the value of $x_{i}$ we need to add multiples of the other equations to the $i^{\text {th }}$ equation, till we only have the $x_{i}$ term left on the left hand side.
To do this we first make the coefficient of $a_{i n}$ zero by subtracting $v_{i n} / v_{n n} y_{n}$. Next we make the coefficient of $a_{i, n-1}$ zero by subtracting $v_{i, n-1} / v_{n-1, n-1} y_{n-1}$ and we need to add $\left(v_{i, n-1} / v_{n-1, n-1}\right)\left(v_{n-1, n} / v_{n n}\right) y_{n}$ to make the coefficient of $a_{i n}$ zero again. Continuing this process will lead to the formula in the lemma.

Corollary 6.30. Applying Lemma 6.29 to the system

$$
\Gamma_{j l}|x|^{2 j} f_{j}=x^{2 l} \Delta_{k}^{l} f,
$$

with $\Gamma_{j l}=\lambda_{n-2 j, j}^{l}, 0 \leq j \leq\lfloor n / 2\rfloor$, gives

$$
|x|^{2 j} f_{j}=\frac{1}{\lambda_{n-2 j, j}^{j}}\left(x^{2 j} \Delta_{k}^{j} f+\sum_{a=j+1}^{\lfloor n / 2\rfloor} \sum_{b=a}^{\lfloor n / 2\rfloor}-\frac{\lambda_{n-2 j, j}^{b}}{\lambda_{n-2 b, b}^{b}}\left(\prod_{c=b+1}^{n}-\frac{\lambda_{n+2-2 c, c-1}^{c}}{\lambda_{n-2 c, c}^{c}}\right) x^{2 b} \Delta_{k}^{b} f\right) .
$$

### 6.3 The adjoint of $T_{u}$

In this section, we will compute the adjoint of $T_{u}, u \in \mathbb{R}^{m}$, on the space $H_{k}\left(\mathbb{R}^{m}\right)$ of all $k$ harmonic polynomials on $\mathbb{R}^{m}$ with the inner product $\langle f, g\rangle_{h}=\int_{S} f g h^{2} d \omega$. We will also look at the operator $\sum_{i=1}^{m} T_{i}^{*} T_{i}$.

Lemma 6.31. 9, Thm 2.4] Let $f \in P_{n}$ arbitrary. Then

$$
\int_{S} \frac{\partial f}{\partial x_{i}} d \omega=(n+m-1) \int_{S} x_{i} f(x) d \omega
$$

Proof. Using polar coordinates, we see that

$$
\begin{equation*}
\int_{|x| \leq 1} g(x) d x=c_{m} \int_{0}^{1} \int_{S} r^{m-1} g(r x) d r d \omega(x) \tag{6.8}
\end{equation*}
$$

for some constant $c_{m}$ independent of $g$ and each continuous function $g$ on the closed unit ball.
Set $g=\partial f / \partial x_{i}\left(1-|x|^{2}\right)$. Since $\partial f / \partial x_{i}$ is homogeneous of degree $n-1$, we can put the $r$-dependence in (6.8) in a different integral which leads to

$$
\begin{aligned}
\int_{S} \frac{\partial f(x)}{\partial x_{i}} d \omega & =A_{1} \int_{|x| \leq 1} \partial f / \partial x_{i}\left(1-|x|^{2}\right) d x \\
& =-A_{1} \int_{|x| \leq 1} f(x)(\partial / \partial x)\left(1-|x|^{2}\right) d x \\
& =2 A_{1} \int_{|x| \leq 1} x_{i} f(x) d x \\
& =2\left(A_{1} / A_{2}\right) \int_{S} x_{i} f(x) d \omega
\end{aligned}
$$

where

$$
A_{1}=\left(c_{m} \int_{0}^{1} r^{m+n-1}\left(1-r^{2}\right) d r\right)^{-1}
$$

and

$$
A_{2}=\left(c_{m} \int_{0}^{1} r^{m+n-1} d r\right)^{-1}
$$

so $\left(2 A_{1} / A_{2}\right)=m+k-1$.
Lemma 6.32. [9, Prop. 2.2] For $f \in C^{2}\left(\mathbb{R}^{m}\right)$ we have that

$$
\begin{equation*}
\Delta_{k}\left(x_{i} f(x)\right)=\left(x_{i} \Delta_{k}+2 T_{i}\right) f(x) \tag{6.9}
\end{equation*}
$$

Proof.

$$
\begin{aligned}
\Delta_{k}\left(x_{i} f(x)\right)= & x_{i} \Delta f(x)+2 \frac{\partial f}{\partial x_{i}}+2 \sum_{\alpha \in R_{+}} k_{\alpha}\left[\frac{x_{i}\langle\alpha, \nabla f\rangle}{\langle\alpha, x\rangle}+\frac{f(x) \alpha_{i}}{\langle\alpha, x\rangle}-\frac{x_{i} f(x)-\left(r_{\alpha}(x)\right)_{i} f\left(r_{\alpha}(x)\right)}{\langle\alpha, x\rangle^{2}}\right] \\
= & x_{i} \Delta f(x)+2 \frac{\partial f}{\partial x_{i}} \\
& +2 \sum_{\alpha \in R_{+}} k_{\alpha}\left[\frac{x_{i}\langle\alpha, \nabla f\rangle}{\langle\alpha, x\rangle}+\frac{f(x) \alpha_{i}}{\langle\alpha, x\rangle}-\frac{x_{i}\left(f(x)-f\left(r_{\alpha} x\right)\right)+\left(x_{i}-\left(r_{\alpha}(x)\right)_{i}\right) f\left(r_{\alpha}(x)\right)}{\langle\alpha, x\rangle^{2}}\right] \\
= & x_{i} \Delta f(x)+2 \frac{\partial f}{\partial x_{i}} \\
& +2 \sum_{\alpha \in R_{+}} k_{\alpha}\left[\frac{x_{i}\langle\alpha, \nabla f\rangle}{\langle\alpha, x\rangle}+\frac{f(x) \alpha_{i}}{\langle\alpha, x\rangle}-\frac{x_{i}\left(f(x)-f\left(r_{\alpha} x\right)\right)}{\langle\alpha, x\rangle^{2}}-\frac{\langle x, \alpha\rangle \alpha_{i} f\left(r_{\alpha} x\right)}{\langle\alpha, x\rangle^{2}}\right] \\
= & x_{i} \Delta_{k}+2 T_{i} f(x)
\end{aligned}
$$

where we have used the product rules and have added and subtracted $\left(x_{i} f\left(r_{\alpha} x\right)\right)$ in the last fraction of the second term.

Lemma 6.33. [9, Prop 2.3] For $f \in H_{k, n}$, we have

$$
x_{i} f-(N+2 n+2 \gamma-2)^{-1}|x|^{2} T_{i} f \in H_{k, n+1} .
$$

Proof. We have for $f \in H_{k, n}$ and $c \in \mathbb{R}$ that

$$
\Delta_{k}\left(x_{i} f-c|x|^{2} T_{i} f\right)=x_{i} \Delta_{k} f+(2-4(n+\gamma-1+m / 2) c) T_{i} f+|x|^{2} \Delta_{h} T_{i} f
$$

by (6.5) and (6.9).
Since $T_{i}$ and $\Delta_{k}$ commute, this expression equals 0 for $c=(N+2 n+2 \gamma-2)^{-1}$. In particular

$$
x_{i} f-(N+2 n+2 \gamma-2)^{-1}|x|^{2} T_{i} f \in H_{k, n+1}
$$

Let $\langle\cdot, \cdot\rangle_{h}$ be the inner product of $L^{2}\left(S, h^{2} d \omega\right)$. Let $H_{k}\left(\mathbb{R}^{m}\right)$ be the space of all harmonic polynomials. Note that each element of $H_{k}\left(\mathbb{R}^{m}\right)$ is uniquely determined by its restriction to $S$, see Corollary 6.24 .

Theorem 6.34. [9, Thm 2.1] The adjoint of $T_{i}$, as operator on $H_{k}\left(\mathbb{R}^{m}\right)$ with the inner product inherited from $\langle\cdot, \cdot\rangle_{h}$, is given by

$$
T_{i}^{*} p(x)=(m+2 n+2 \gamma)\left(x_{i} p(x)-(m+2 n+2 \gamma-2)^{-1}|x|^{2} T_{i} p(x)\right),
$$

for $p \in H_{k, n}$. Here $\gamma=\operatorname{deg}(h)$ as in the previous section.
Proof. Let $f \in H_{k, n+1}$ and $g \in H_{k, n}$. Then

$$
\begin{aligned}
\int_{S}\left(f T_{i} g+g T_{i} f\right) h^{2} d \omega= & \int_{S} f \partial_{i} g+g \partial_{i} f+2 \sum_{\alpha \in R_{+}} k_{\alpha} \alpha_{i} \frac{f(x) g(x)}{\langle\alpha, x\rangle} h(x)^{2} d \omega(x) \\
& -\int_{S} \sum_{\alpha \in R_{+}} k_{\alpha} \alpha_{i} \frac{f\left(r_{\alpha} x\right) g(x)+f(x) g\left(r_{\alpha} x\right)}{\langle\alpha, x\rangle} h(x)^{2} d \omega(x) .
\end{aligned}
$$

The first integral is equal to

$$
\int_{S} \partial_{i}\left(f(x) g(x) h(x)^{2}\right) d \omega
$$

because

$$
\partial_{i} h^{2}(x)=\partial_{i}\left(\prod_{\alpha \in R_{+}}|\langle x, \alpha\rangle|^{2 k_{\alpha}}\right)=2 \sum_{\alpha \in R_{+}} k_{\alpha} \frac{\alpha_{i}}{\langle x, \alpha\rangle} h(x)^{2} .
$$

The second integral is equal to 0 , since

$$
\begin{aligned}
\int_{S} \frac{f\left(r_{\alpha} x\right) g(x)}{\langle x, \alpha\rangle} h^{2}(x) d \omega(x) & =\int_{S} \frac{f(x) g\left(r_{\alpha} x\right)}{\left\langle r_{\alpha} x, \alpha\right\rangle} h^{2}\left(r_{\alpha} x\right) d \omega(x) \\
& =-\int_{S} \frac{f(x) g\left(r_{\alpha} x\right)}{\langle x, \alpha\rangle} h^{2}(x) d \omega(x),
\end{aligned}
$$

because $h(x)$ is $G$-invariant and $r_{\alpha}(\alpha)=-\alpha$.
So

$$
\int_{S} \partial_{i}\left(f(x) g(x) h^{2}(x)\right) d \omega(x)=\int_{S}\left(f(x) T_{i} g(x)+g(x) T_{i} f(x)\right) h^{2}(x) d \omega(x) .
$$

By Lemma 6.31 we have

$$
\int_{S} \partial_{i}\left(f(x) g(x) h^{2}(x)\right) d \omega(x)=(2 \gamma+2 n+m) \int_{S} x_{i} f(x) g(x) h^{2}(x) d \omega(x)
$$

and together these identities lead to

$$
\int_{S} T_{i}(f) g h^{2} d \omega=\int_{S} f\left((2 n+2 \gamma+m) x_{i} g\right) h^{2} d \omega-\int_{S} T_{i}(g) f h^{2} d \omega .
$$

The integral $\int_{S} T_{i}(g) f h^{2} d \omega$ equals 0 , because $T_{i} g \in H_{k, n-1}, f \in H_{k, n+1}$ and $H_{k, n-1} \perp H_{k, n+1}$. Finally, by Lemma 6.33 the function

$$
g_{i}(x)=(m+2 n+2 \gamma)\left(x_{i} g(x)-(m+2 n+2 \gamma)^{-1}\left(T_{i} g\right)(x)\right)
$$

is an element of $H_{k, n+1}$ that satisfies

$$
\int_{S} f g_{i} h^{2} d \omega=\int_{S} T_{i}(f) g h^{2} d \omega
$$

so we have found the adjoint of $T_{i}$ on $H_{k}\left(\mathbb{R}^{m}\right)$, with the inner product $\langle\cdot, \cdot\rangle_{h}$.
Lemma 6.35. [9, Prop. 2.5] Let $f \in H_{k, n}$. The selfadjoint operator $\sum_{i=1}^{m} T_{i}^{*} T_{i}: H_{k, n} \rightarrow P_{n}$ satisfies

$$
\sum_{i=1}^{m} T_{i}^{*} T_{i} f=(2 n+2 \gamma-2) \sum_{i=1}^{m} x_{i} T_{i} f .
$$

Also

$$
\begin{equation*}
\sum_{i=1}^{m} x_{i} T_{i} f=n f+\sum_{\alpha \in R_{+}} k_{\alpha}\left(f(x)-f\left(r_{\alpha} x\right)\right) . \tag{6.10}
\end{equation*}
$$

Proof.

$$
\begin{aligned}
\sum_{i=1}^{m} T_{i}^{*} T_{i} f & =\sum_{i=1}^{m}(m+2 n+2 \gamma-2)\left(x_{i} T_{i} f(x)-(m+2 n+2 \gamma-2)^{-1}|x|^{2} T_{i}^{2} f(x)\right) \\
& =\sum_{i=1}^{m}(m+2 n+2 \gamma-2) x_{i} T_{i} f(x)
\end{aligned}
$$

because $f(x)$ is $k$-harmonic.
The second statement follows from the definition of the Dunkl operator as given in Definition 6.3 .

The operators $\sum_{i=1}^{m} T_{i}^{*} T_{i}$ and $\sum_{i=1}^{m} x_{i} T_{i}$ are homogeneous of degree 0 . Their eigenvalues and eigenfunctions contain a lot of information about Dunkl operators. As we will show in 7.3, if $\sum_{i=1}^{m} T_{i}^{*} T_{i}$ has a zero eigenvalue, either $(m+2 n+2 \gamma-2)=0$ or an eigenvalue of $\sum_{i=1}^{m} x_{i} T_{i}$ equals 0 . In the first case the Fischer decomposition in Theorem 6.16 breaks down. In the other case the future construction of intertwining operator in Theorem 7.14 breaks down. In the next section we will show that all eigenvalues of $\sum_{i=1}^{m} x_{i} T_{i}$ are positive for $k>0$, so the eigenvalues of $\sum_{i=1}^{m} T_{i}^{*} T_{i}$ are also positive for $k>0$.

### 6.4 The group algebra

In this section we will have a look of the $\mathbb{C}$-valued group algebra $\mathbb{C} G$, which is related to the Weyl group $G$. We will construct the Fourier transform associated with the conjugation invariant functions $G \rightarrow \mathbb{C}$. We will use the group algebra to find the eigenvalues of operator $\sum_{i=1}^{m} x_{i} T_{i}: P \rightarrow P$.
Definition 6.36. For a finite group $G$ and a field $\mathbb{K}$, the group algebra $\mathbb{K} G$ is a $\mathbb{K}$-linear space with basis $G$. The multiplication on $\mathbb{K} G$ is the bilinear map $\mathbb{K} G \times \mathbb{K} G \rightarrow \mathbb{K} G$ given by $g \cdot h=g h$ for all $g, h \in G$. Thus

$$
\sum_{g \in G} c_{g} g \cdot \sum_{h \in G} d_{h} h=\sum_{j \in G} e_{j} j,
$$

with

$$
e_{k}=\sum_{\substack{g, h \in G \\ g h=k}} c_{g} d_{h} .
$$

The group algebra is characterized up to isomorphisms by the following universal property. For any map $\phi: G \rightarrow A$ into an associative $\mathbb{K}$-algebra, such that $\phi(g h)=\phi(g) \phi(h)$, there is a unique algebra homomorphism $\bar{\phi}: \mathbb{K} G \rightarrow A$ such that $\phi=\bar{\phi} \circ i$, where is the inclusion of $G$ into $\mathbb{K} G$.
The left regular representation of $G$ on $P\left(\mathbb{R}^{m}\right)$, defined by

$$
L(w) f(x)=f\left(w^{-1} x\right)
$$

for $w \in G$ and $f \in P\left(\mathbb{R}^{m}\right)$, can be extended to a representation of the group algebra $\mathbb{C} G$, by

$$
L\left(\sum_{w \in G} c_{w} w\right) f(x)=\sum_{w \in G} c_{w} f\left(w^{-1} x\right)
$$

Note that $L(c) P_{n} \subset P_{n}$ for all $c \in \mathbb{C} G, n \in \mathbb{N}$. The representation $\left.L\right|_{P_{n}}$ is finite dimensional and can be decomposed into irreducible components and so the representation $L$ can be decomposed into irreducible homogeneous components. Each irreducible component is an irreducible $\mathbb{C} G$-module.

Definition 6.37. [9, p.176] Define $\phi \in \mathbb{C} G$ by

$$
\phi=\sum_{\alpha \in R_{+}} k_{\alpha}\left(1-r_{\alpha}\right) .
$$

By using the left regular representation of $\mathbb{C} G$ on $C^{1}\left(\mathbb{R}^{m}\right)$, we find

$$
L(\phi)(f)=\sum_{\alpha \in R_{+}} k_{\alpha}\left(f(x)-f\left(r_{\alpha} x\right)\right)=\sum_{i=1}^{m}\left(x_{i} T_{i}-x_{i} \partial_{i}\right) f,
$$

for $f \in C^{1}\left(\mathbb{R}^{m}\right)$.
Lemma 6.38. [9, p.176] The element $\phi$ is a central element of the group algebra of $G$.
Proof. We can write $\phi=\frac{1}{2} \sum_{\alpha \in R} k_{\alpha}\left(1-r_{\alpha}\right)$. Then for all $g \in G$ we have

$$
\begin{aligned}
g \phi g^{-1} & =\frac{1}{2} \sum_{\alpha \in R} k_{\alpha}\left(1-r_{g \cdot \alpha}\right) \\
& =\frac{1}{2} \sum_{\alpha \in R} k_{g^{-1} \alpha}\left(1-r_{\alpha}\right) \\
& =\frac{1}{2} \sum_{\alpha \in R} k_{\alpha}\left(1-r_{\alpha}\right) \\
& =\phi,
\end{aligned}
$$

because $k_{\alpha}$ is $G$-invariant.
Definition 6.39. Let $V$ be an $l$-dimensional irreducible component of $\mathbb{C} G$, with associated representation $\rho$. Denote by $\chi$ the character of $\rho$, which is given by $\chi(w)=\operatorname{tr}(\rho(w))$.
Denote the set of all characters of $G$ by $\hat{G}$,
Note that the trace of $\rho$ is well-defined, because $\rho$ is finite-dimensional linear map between vector spaces.

Definition 6.40. For $c \in \mathbb{C} G$ define the map $M_{c}: \mathbb{C} G \rightarrow \mathbb{C} G$ by $M_{c}(d)=c d$.
Lemma 6.41. [10, p. 109] The eigenvalues of $M_{\phi}$ on the group algebra are given by

$$
\begin{equation*}
\lambda(\chi)=\sum_{\alpha \in R_{+}} k_{\alpha}\left(1-\chi\left(r_{\alpha}\right) / \chi(1)\right), \tag{6.11}
\end{equation*}
$$

for $\chi \in \hat{G}$. Let $V$ be an irreducible component of $\mathbb{C} G$, with character $\chi_{V}$. Then $f \in V$ is an eigenfunction of $\phi$, with eigenvalue $\lambda\left(\chi_{V}\right)$.

Proof. Let $V$ be an irreducible component of $\mathbb{C} G$, with character $\chi_{V}$. By Schur's lemma the element $\phi$ acts as a multiple of the identity on $V$, so $\left.M_{\phi}\right|_{V}=\lambda I$. Hence

$$
\begin{aligned}
\lambda \chi_{V}(1) & =\lambda \operatorname{dim}(V) \\
& =\operatorname{tr}\left(M_{\phi} \mid V\right) \\
& =\operatorname{tr}\left(\sum_{\alpha \in R_{+}} k_{\alpha}\left(I_{V}-M_{r_{\alpha}} \mid V\right)\right) \\
& =\sum_{\alpha \in R_{+}} k_{\alpha} \operatorname{dim}(V)-\sum_{\alpha \in R_{+}} k_{\alpha} \operatorname{tr}\left(\left.M_{r_{\alpha}}\right|_{V}\right) \\
& =\sum_{\alpha \in R_{+}} k_{\alpha}\left(\chi_{V}(1)-\chi_{V}\left(r_{\alpha}\right)\right),
\end{aligned}
$$

which implies that

$$
\lambda=\sum_{\alpha \in R_{+}} k_{\alpha}\left(1-\chi_{V}\left(r_{\alpha}\right) / \chi_{V}(1)\right) .
$$

This shows that the eigenvalues of $\phi$ are given by

$$
\lambda(\chi)=\sum_{\alpha \in R_{+}} k_{\alpha}\left(1-\chi\left(r_{\alpha}\right) / \chi(1),\right.
$$

for $\chi \in \hat{G}$.
Also the identity $\left.M_{\phi}\right|_{V}=\lambda I$, shows that $f \in V$ is an eigenfunction of $\phi$ and the eigenvalue was computed to be $\lambda\left(\chi_{V}\right)$.
Corollary 6.42. Let $c=\sum_{w \in G} c_{w} \in \mathcal{Z} \mathbb{C} G$. Then the eigenvalues of $M_{c}$ on the group algebra are given by

$$
\begin{equation*}
\lambda_{c}(\chi)=\sum_{w \in G} c_{w} \chi(w) / \chi(1) \tag{6.12}
\end{equation*}
$$

for $\chi \in \hat{G}$. Let $V$ be an irreducible component of $\mathbb{C} G$, with character $\chi_{V}$. Then $f \in V$ is an eigenfunction of $c$, with eigenvalue $\lambda_{c}\left(\chi_{V}\right)$.
Proof. We can prove this by replacing $\phi$ with $c$ in the proof of Lemma 6.41.
Corollary 6.43. Consider the representation of $\mathbb{C} G$ on $P\left(\mathbb{R}^{m}\right)$. Let $V$ be a irreducible component of dimension l, contained in $P_{n}\left(\mathbb{R}^{m}\right)$, for some $n$. Let $\chi$ be the associated character. Then $\sum_{i=1}^{m} x_{i} T_{i}$ acts as a scalar on $V$. This scalar is given by

$$
\sum_{\alpha \in R_{+}} k_{\alpha}\left(1-\chi\left(r_{\alpha}\right) / \chi(1)\right)+n
$$

Proof. This follows from equations (6.10) and (6.11).
Suppose the group $G$ has $j$ conjugacy classes of reflections, each of which we can write as $\left\{\alpha_{i, j}, \ldots, \alpha_{i, m_{i}}\right\}$, for $1 \leq i \leq j$ and $m_{i} \in \mathbb{N}$. If $k_{i}$ is the common value of $k_{\alpha}$ on the $i^{t h}$ conjugacy class, the eigenvalues of $\phi$ are given by

$$
\lambda(\chi)=\sum_{i=1}^{l} m_{i} k_{i}\left(1-\chi\left(\alpha_{i, 1}\right) / \chi(1)\right),
$$

for any irreducible character $\chi$ of $G$ (see [10]).
For an irreducible character $\chi \in \hat{G}$, and for $1 \leq i \leq l$, the number $m_{i} \chi\left(\alpha_{i, 1}\right) / \chi(1) \in \mathbb{Z}$ (see [10, p.110]).

Lemma 6.44. 10, Cor. 2.2] For an irreducible character $\chi \in \hat{G}$, we have that $\lambda(\chi)=$ $\sum_{i=1}^{l} k_{i} n_{i}$, with $n_{i} \in \mathbb{Z}$ and $0 \leq n_{i} \leq 2 m_{i}$. For the trivial character we find $n_{i}=0$. Let $\rho: G \rightarrow\{-1,1\}$ be the unique representation of $G$, with $r_{\alpha}=-1$, for all $\alpha \in R$. For the character $\chi_{\rho}$ we find $n_{i}=2$.

Proof. Because $m_{i} \chi\left(\alpha_{i, 1}\right) / \chi(1) \in \mathbb{Z}$ and $m_{i} \in \mathbb{Z}$, we have that $n_{i}=m_{i}\left(\chi(1)-\chi\left(\alpha_{i, 1}\right)\right) / \chi(1) \in$ $\mathbb{Z}$. The inequality $0 \leq n_{i} \leq m_{i}$ follows from the inequality $|\chi(w)| \leq \chi(1), w \in G$.
For the trivial character, we have that $\chi\left(r_{\alpha}\right)=1$, so $n_{i}=0$, for all $i$.
For the character $\chi_{\rho}$, we have that $\chi\left(r_{\alpha}\right)=-1$, so $n_{i}=2$, for all $i$.
Let $C(G$, class $)$ be the space of conjugation invariant functions $G \rightarrow \mathbb{C}$. We have the inner product $\langle\cdot, \cdot\rangle_{G}: C(G$, class $) \times C(G$, class $) \rightarrow \mathbb{R}$, given by

$$
\langle f, g\rangle_{G}=1 /|G| \sum_{w \in G} f(w) \overline{g(w)}
$$

The set $\hat{G}$ of all character on $G$ is an orthonormal basis of $C$ ( $G$, class).
Let $\sum_{w \in G} c_{w} w \in \mathcal{Z} \mathbb{C} G$, then $f: w \rightarrow c_{w}$ is a class function. Conversely, if $f \in C(G$, class $)$, then $c=\sum_{w \in G} f(w) w \in \mathcal{Z} \mathbb{C} G$. Thus $\mathcal{Z} \mathbb{C} G \simeq C(G$, class $)$.

The Fourier transform is a linear map $C(G$, class $) \rightarrow C(\widehat{G})$, defined by

$$
\hat{f}(\chi)=\langle f, \chi\rangle=\frac{1}{|G|} \sum_{w \in G} f(w) \overline{\chi(w)} .
$$

The Fourier transform of $c \in \mathcal{Z} \mathbb{C} G$ is defined by

$$
\hat{c}(\chi)=\frac{1}{|G|} \sum_{w \in G} c_{w} \overline{\chi(w)} .
$$

The Fourier inverse transform is the map $\mathscr{F}^{-1}: C(\widehat{G}) \rightarrow C(G$, class $)$ defined by

$$
\mathscr{F}^{-1}(F)=\sum_{\chi \in \widehat{G}} F(\chi) \chi .
$$

The Fourier inversion formula is given by $f=\sum_{w \in \widehat{G}} \hat{f}(\chi) \chi$. For $c \in \mathcal{Z} \mathbb{C} G$ this means

$$
c_{w}=\sum_{\chi \in \widehat{G}} \hat{c}(\chi) \chi(w) .
$$

For $\chi \in \widehat{G}$, define the map $\psi_{\chi}: \mathcal{Z C} G \rightarrow C(\widehat{G})$ by

$$
\psi_{\chi}(c)=\sum_{w \in G} c_{w} \chi(w) / \chi(1) .
$$

Denote by $\check{\chi}(w)=\chi\left(w^{-1}\right)$ the character of the dual representation. Since $\{\check{\chi}\}_{\chi \in \widehat{G}}$ is an orthonormal basis of $\mathcal{Z C} G \simeq C(G$, class $)$, we find

$$
\begin{aligned}
c_{w} & =\sum_{\chi \in \widehat{G}}\langle c, \check{\chi}\rangle \check{\chi}(w) \\
& =\sum_{\chi \in \widehat{G}} \sum_{z \in G} c_{z} \overline{\chi\left(z^{-1}\right)} \chi\left(w^{-1}\right) \\
& =\sum_{\chi \in \widehat{G}} \sum_{z \in G} \frac{\chi(1)}{|G|} \psi_{\chi}(c) \overline{\chi(w)},
\end{aligned}
$$

for $\sum_{w \in G} c_{w} w \in \mathcal{Z} \mathbb{C} G$. Note that $\psi_{\chi}(c)$ is closely related to the fourier transform, because

$$
\psi_{\chi}(c)=\frac{1}{\chi(1)} \sum_{w \in G} c_{w} \overline{\chi\left(w^{-1}\right)}=\frac{1}{\check{\chi}(1)} \sum_{w \in G} c_{w} \bar{\chi}(w)=\frac{|G|}{\hat{\chi}(1)} \hat{c}(\check{\chi}) .
$$

Let $c, d \in \mathcal{Z} \mathbb{C} G$. By 6.12 the eigenvalues of multiplication by $c$ are given by $\psi_{\chi}(c), \chi \in \hat{G}$ and the eigenvalues of multiplication by $d$ are given by $\psi_{\chi}(d), \chi \in \hat{G}$.
Let $V$ be an irreducible component of the left regular representation of $\mathbb{C} G$. Let $\chi$ be the associated character. We have that $L(c) L(d) f=L(c d) f$ for $f \in V$, because $L$ is a representation. By Corollary 6.43, we have that $L(c) L(d) f=\psi_{\chi}(c) \psi_{\chi}(d) f$ and $L(c d) f=\psi_{\chi}(c d) f$ for $f \in V$.
By combining this, we see that $\psi_{\chi}(c) \psi_{\chi}(d)=\psi_{\chi}(c d)$, for all $\chi \in \hat{G}$ and for all $c, d \in \mathcal{Z} \mathbb{C} G$, so $\phi_{\chi}$ is a algebra homomorphism $\mathcal{Z} \mathbb{C} G \rightarrow \mathbb{C}$.
Next let $c=\exp ((\log t) \phi)=\exp \left((\log t) \sum_{\alpha \in R_{+}} k_{\alpha}\left(1-r_{\alpha}\right)\right)$, where $\phi$ was defined in Definition 6.37 and the exponent of $\phi$ is evaluated as a power series in $\mathbb{C} G$. The element $c$ is central in $\mathbb{C} G$, because $\phi$ is a central element of $\mathbb{C} G$. Applying the transform $\psi_{\chi}$ to $c$ yields

$$
\begin{aligned}
\psi_{\chi}(c) & =\exp \left((\log t) \sum_{\alpha \in R_{+}} k_{\alpha}\left(\psi_{\chi}(1)-\psi_{\chi}\left(r_{\alpha}\right)\right)\right) \\
& =\exp \left((\log t) \sum_{\alpha \in R_{+}} k_{\alpha}\left(1-\chi\left(r_{\alpha}\right) / \chi(1)\right)\right) \\
& =t^{\lambda(\chi)}
\end{aligned}
$$

where it is used that $\psi_{\chi}$ is a homomorphism. The constants $\lambda(\chi)$ are defined in 6.11.
Definition 6.45. [10, p.111] Let $R$ be a root system with Weyl group $G$ and weight function $k$. Let $\phi$ as in Definition 6.37. The element $\phi$ is an element of $\mathbb{C} G$, so we can use it to define coefficients $p_{w}(t)$ by

$$
\exp \left((\log t) \sum_{\alpha \in R_{+}} k_{\alpha}\left(1-r_{\alpha}\right)\right)=\frac{1}{|G|} \sum_{w \in G} p_{w}(t) w \in \mathbb{C} G
$$

for $0 \leq t \leq 1$, where the exponent is evaluated as a power series in $\mathbb{C} G$.

By the Fourier inversion formula on $\mathcal{Z C} G$, we have that

$$
\begin{equation*}
p_{w}(t)=\sum_{\chi \in \hat{G}} \chi(1) \bar{\chi}(w) t^{\lambda(\chi)} . \tag{6.13}
\end{equation*}
$$

For Coxeter groups we have a stronger statement.
Theorem 6.46. [10, Thm 3.1] For any Coxeter group $G$, and $w \in G$, the coefficients of $t^{\lambda(\chi)}$ in $p_{w}(t)$ are integers. Also $\overline{\chi(w)}=\chi(w)$.

Proof. See [10, p. 112] for a proof by checking the theorem in character tables.

## Chapter 7

## The intertwining operator

In this chapter we will construct the operator which intertwines the actions of the partial differential operators and the Dunkl operators.
The main result of this chapter is Theorem 7.14, which is stated as
Main Theorem. Let $P=P\left(\mathbb{R}^{m}\right)$. There exists a unique linear operator $V: P \rightarrow P$, such that $V(1)=1, V P_{n} \subset P_{n}$ and $T_{i} V f=V\left(\partial_{i} f\right), 1 \leq i \leq m$. The operators $T_{i}$ were defined in Definition 6.3. The operator $V$ is invertible.

This operator is called an intertwining operator, since it intertwines the action of the Dunkl operators with the action of the partial derivatives. See (7.8) for the precise form of $V$.
A proof of this theorem was given by Dunkl in [10, p. 111-116]. In this chapter we will give some motivation for this proof and we will look at some properties of the intertwining operator.
In Section 7.3 we shall have a short look at negative weight functions and show that the intertwining operator might not exist for certain negative weight functions.

First consider that we have found the intertwining operator $V$. Then it is useful to have a look at the image of the monomials under $V$.
Let $x^{\alpha}$ be a monomial then $\partial_{i} x^{\alpha}=\alpha_{i} x^{\alpha-e_{i}}=\frac{\alpha!}{\left(\alpha-e_{i}\right)!} x^{\alpha-e_{i}}$, where we have used multi-index notation and view $\alpha$ as a vector. This can be generalized to $\partial^{\alpha} x^{\beta}=\frac{\alpha!}{(\alpha-\beta)!} x^{\alpha-\beta}$, for $\alpha \leq \beta$ and $\partial^{\alpha} x^{\beta}=0$, for $\alpha>\beta$. Here $<$ denotes the partial order defined by $\alpha<\beta$ if $\alpha_{i} \leq \beta_{i}$ for $1 \leq i \leq m$ and $\alpha_{i}<\beta_{i}$ for at least one $i$.
The monomials do not behave in this way under the action of Dunkl operators. However, the relations above yield the idea of trying to create Dunkl monomials $y_{\alpha} \in P$ in an inductive way, such that $y_{0}=1$ and

$$
T^{\alpha} y_{\beta}=\frac{\alpha!}{(\beta-\alpha)!} y_{\beta-\alpha} .
$$

Then the intertwining operator should act by $V x^{\alpha}=y_{\alpha}$, since then $V \partial_{\beta} x^{\alpha}=V(\alpha-\beta)!x^{\alpha-\beta}=$ $(\alpha-\beta)!y_{\alpha-\beta}=T^{\beta} y_{\alpha}=T^{\beta} V x_{\alpha}$. Note that the Dunkl-monomials do not obey rules like $y_{\alpha} y_{\beta}=y_{\alpha+\beta}$.

Definition 7.1. Let $f_{1}, \ldots f_{m}$ be an m-tuple of $C^{1}$-functions $\mathbb{R}^{m} \rightarrow \mathbb{R}$. The $m$-tuple is called exact if $\partial_{i} f_{j}=\partial_{j} f_{i}$, for all $i, j$. The $m$-tuple is called k-exact if $T_{i} f_{j}=T_{j} f_{i}$, for all $i, j$. Denote the space of all $k$-exact tuples by $\Omega^{k}$ and denote the space of all exact tuples by $\Omega$.

Also denote the space of all $k$-exact tuples which are homogeneous polynomials of degree $n$ by $\Omega_{n}^{k}$ and denote the space of exact tuples which are homogeneous polynomials of degree $n$ by $\Omega_{n}$.

We can view $\nabla$ as operator from $C^{2}\left(\mathbb{R}^{m}\right)$ to $\Omega$ and we can view the operator $\left.\nabla\right|_{P_{n}}$ as operator from $P_{n}$ to $\Omega_{n-1}$.
Definition 7.2. Define $W: \Omega \rightarrow C^{2}\left(\mathbb{R}^{m}\right)$ by

$$
(W f)(x)=\int_{0}^{1}\langle x, f(t x)\rangle d t .
$$

Lemma 7.3. Let $f \in \Omega$. Then $\partial_{i} W f=f_{i}$.
Proof. Let $f=\left(f_{1}, \ldots f_{m}\right) \in \Omega$ arbitrary, then

$$
\begin{aligned}
\frac{\partial(W f)}{\partial x_{i}}(x) & =\int_{0}^{1}\left(f_{i}(t x)+\sum_{j=1}^{m} t x_{j} \frac{\partial f_{j}(t x)}{\partial x_{i}}\right) d t \\
& =\int_{0}^{1}\left(f_{i}(t x)+\sum_{j=1}^{m} t x_{j} \frac{\partial f_{i}(t x)}{\partial x_{j}}\right) d t \\
& =\int_{0}^{1}\left(f_{i}(t x)+t \frac{d}{d t} f_{i}(t x)\right) d t \\
& =\int_{0}^{1} \frac{d}{d t}\left(t f_{i}(t x)\right) d t \\
& =f_{i}(x)
\end{aligned}
$$

which proves the lemma.
Corollary 7.4. The operator $W: \Omega_{n} \rightarrow P_{n+1}$, can be seen as the two sided inverse of $\nabla$ restricted to $P_{n+1}$. Also the operator $W: \Omega \rightarrow C^{2}\left(\mathbb{R}^{m}\right)$, can be seen as the two sided inverse of $\nabla$ restricted to $C^{2}\left(\mathbb{R}^{m}\right)$.

Proof. By Lemma 7.3, $\nabla(W f)(x)=f(x)$, for $f \in \Omega_{n}$. Let $F \in P_{n+1}$. Then $\nabla F$ is a exact $m$-tuple. By Lemma $7.3 \nabla(W(\nabla F))=\nabla F$, so $W(\nabla F)(x)-F(x)=F(0)$. Since $F$ is a homogeneous polynomial and $W \nabla F$ is a homogeneous polynomial $F=W \nabla F$.
The second statement follows in a trivial way, since in this case $\nabla F$ is differentiable, which shows that $F \in C^{2}\left(\mathbb{R}^{m}\right)$.

By the intertwining property, the operator $V$ must map exact $m$-tuples of functions into $k$ exact $m$-tuples of functions. Our goal is to generalize Lemma 7.3 in the right way to find an inverse $W_{k}: \Omega^{k} \rightarrow C^{2}\left(\mathbb{R}^{m}\right)$ of $\nabla_{k}: C^{2}\left(\mathbb{R}^{m}\right) \rightarrow \Omega^{k}$. This gives the relation $V F=W_{k} V \nabla F$, for all $F \in C\left(\mathbb{R}^{m}\right)$.
Consider $F \in P_{n}$, then $V F \in P_{n}$. Since $\Omega_{k, n}^{k} \subset P_{n}^{m}$, we have that $(V \nabla F)_{i} \in P_{n-1}, 1 \leq i \leq m$. Because $V 1=1$, we can use these relations to define the intertwining operator in an inductive manner on $P\left(\mathbb{R}^{m}\right)$ by $V_{n+1}=W_{k} V_{n} \nabla F$, where $V_{n}=\left.V\right|_{P_{n}}$.
Of course we want to extend this inductive definition of $V$ to all of $C^{1}\left(\mathbb{R}^{m}\right)$, but so far this has only be done for the root system $A_{1}$ (see [11, Thm 5.1]).

We try the Ansatz

$$
\left(W_{k} f\right)(x)=\frac{1}{|G|} \sum_{w \in G} \int_{0}^{1} q_{w}(t)\langle w x, f(t w x)\rangle d t,
$$

where $q_{w}:(0,1] \rightarrow \mathbb{R}$ are differentiable, as generalization of Lemma 7.3 . In the following section, we will show that the correct generalization is indeed of this form and we shall compute the functions $q_{w}(t)$.

### 7.1 Construction of the functions $q_{w}(t)$

We start by looking at the effect of $\nabla_{k}$ on $\langle x, f(t x)\rangle$ for a $k$-exact $m$-tuple $f$. Note that $f \in \Omega^{k}$ can be seen as a vector $f=\sum_{i=1}^{m} f_{i} e_{i}$, where $e_{i}$ is the $i^{\text {th }}$ basis vector, so for $w \in G$, we have $w f(x)=\sum_{i, j=1}^{m} w_{i j} e_{j} f_{i}$, where $w_{i j}$ is the matrix of the rotation.

Lemma 7.5. [10, Lemma 3.6] Let $f$ be a $k$-exact m-tuple of $C^{1}$-functions. Then

$$
\nabla_{k}(\langle x, f(t x)\rangle)=f(t x)+t \frac{\partial}{\partial t} f(t x)+\sum_{\alpha \in R_{+}} k_{\alpha}\left(f(t x)-r_{\alpha}\left(f\left(t r_{\alpha} x\right)\right)\right)
$$

for $0<t<1$.
Proof. We will prove this component wise. Choose $l$, with $1 \leq l \leq m$. Then

$$
\begin{align*}
T_{l}\langle x, f(t x)\rangle= & f_{l}(t x)+\sum_{i=1}^{m} x_{i} t \frac{\partial f_{i}}{\partial x_{l}}(t x) \\
& +\sum_{\alpha \in R_{+}} k_{\alpha}\left(\sum_{i=1}^{m} x_{i} f_{i}(t x)-\sum_{i=1}^{m}\left(r_{\alpha} x\right)_{i} f_{i}\left(t r_{\alpha} x\right)\right) \alpha_{l} /\langle x, \alpha\rangle \\
= & f_{l}(t x)+\sum_{i=1}^{m} x_{i} t \frac{\partial f_{i}}{\partial x_{l}}(t x) \\
& +\sum_{\alpha \in R_{+}} k_{\alpha}\left(\sum_{i=1}^{m} x_{i} f_{i}(t x)-\sum_{i=1}^{m}(x-\langle\alpha, x\rangle \alpha)_{i} f_{i}\left(t r_{\alpha} x\right)\right) \alpha_{l} /\langle x, \alpha\rangle \\
= & f_{l}(t x)+\sum_{\alpha \in R_{+}} k_{\alpha}\left\langle f\left(t r_{\alpha} x\right), \alpha\right\rangle \alpha_{l} \\
& +\sum_{i=1}^{m} x_{i} t\left(\frac{\partial f_{i}}{\partial x_{l}}(t x)+\sum_{\alpha \in R_{+}} k_{\alpha}\left(f_{i}(t x)-f_{i}\left(t r_{\alpha} x\right)\right) \alpha_{l} /\langle t x, \alpha\rangle\right), \tag{7.1}
\end{align*}
$$

where we have used that $r_{\alpha}(x)=x-\langle x, \alpha\rangle \alpha$ in the second step. Note that $\left\langle f\left(\operatorname{tr}_{\alpha} x\right), \alpha\right\rangle \alpha=$ $f\left(\operatorname{tr}_{\alpha} x\right)-r_{\alpha}\left(f\left(\operatorname{tr}_{\alpha} x\right)\right)$. Also note that the coefficient of $x_{i} t$ in 7.1) is equal to $T_{l} f_{i}$, which is equal to $T_{i} f_{l}$ by the $k$-exactness of $f$. Applying these results to (7.1) gives

$$
\begin{aligned}
& T_{l}(\langle x, f(t x)\rangle)= \\
& \quad f_{l}(t x)+\sum_{i=1}^{m} x_{i} t \frac{\partial}{\partial x_{i}} f_{l}(t x)+\sum_{\alpha \in R_{+}} k_{\alpha}\left(f_{l}\left(t r_{\alpha} x\right)-r_{\alpha} f\left(t r_{\alpha}(x)\right)_{l}\right) \\
& \quad+\sum_{\alpha \in R_{+}} \sum_{i=1}^{m} \alpha_{i} x_{i} t k_{\alpha}\left(f_{l}(t x)-f_{l}\left(t r_{\alpha} x\right)\right) /\langle t x, \alpha\rangle \\
& =f_{l}(t x)+t \frac{\partial}{\partial t} f_{l}(t x)+\sum_{\alpha \in R_{+}} k_{\alpha}\left(f_{l}(t x)-r_{\alpha}\left(f\left(t r_{\alpha} x\right)\right)_{l}\right),
\end{aligned}
$$

which is the $l^{\text {th }}$ component of the identity in the lemma.
Corollary 7.6. [10, Cor. 3.7] For $w \in G$,

$$
\nabla_{k}(\langle w x, f(t w x)\rangle)=w f(t x w)+t \frac{\partial}{\partial t} w f(t w x)+\sum_{\alpha \in R_{+}} k_{\alpha}\left(w f(t w x)-w r_{\alpha} f\left(t r_{\alpha} w x\right)\right) .
$$

Proof. Recall from Lemma 6.6 the relation

$$
\left(\nabla_{k} L\left(w^{-1}\right) g\right)(x)=w \nabla_{k} g(w x) .
$$

We can apply this result to $g(x)=\langle x, f(x)\rangle$ and use Lemma 7.5 to prove the corollary. By using Corollary 7.6, we can compute the Dunkl gradient of the Ansatz, which gives

$$
\begin{align*}
\left(\nabla_{k} W f\right)(x)= & \frac{1}{|G|} \int_{0}^{1} \sum_{w \in G} q_{w}(t) \nabla_{k}\langle w x, f(t w x)\rangle d t  \tag{7.2}\\
= & \frac{1}{|G|} \sum_{w \in G} \int_{0}^{1}\left(q_{w}(t) w f(t x w)+q_{w}(t) t \frac{\partial}{\partial t}(w f(t w x))\right. \\
& \left.+q_{w}(t) \sum_{\alpha \in R_{+}} k_{\alpha}\left(w f(t w x)-w r_{\alpha} f\left(t r_{\alpha} w x\right)\right)\right) d t \\
= & \frac{1}{|G|} \sum_{w \in G} \int_{0}^{1}\left(\left(q_{w}(t)+q_{w}(t) t \frac{\partial}{\partial t}+\sum_{\alpha \in R_{+}} k_{\alpha}\left[q_{w}(t)-q_{r_{\alpha} w}(t)\right]\right) w f(t w x) d t,\right.
\end{align*}
$$

where we have changed the summation index in the last term.

Lemma 7.7. For $q_{w}(t):(0,1] \rightarrow \mathbb{R}$ differentiable for all $w \in G$, we have

$$
\begin{equation*}
\frac{1}{|G|} \sum_{w \in G}\left(\frac{d}{d t} t q_{w}(t) w f(w x t)\right)=\frac{1}{|G|} \sum_{w \in G}\left(q_{w}(t)+q_{w}(t) t \frac{\partial}{\partial t}+t q_{w(t)}^{\prime}\right) w f(w x t) . \tag{7.3}
\end{equation*}
$$

We also find that

$$
\begin{equation*}
\frac{1}{|G|} \int_{0}^{1} \sum_{w \in G}\left(\frac{d}{d t} t q_{w}(t) w f(w x t)\right) d t=\frac{1}{|G|} \sum_{w \in G} q_{w}(1) f(w x) . \tag{7.4}
\end{equation*}
$$

Proof. The first statement follows by a direct computation using the product rule. The second statement follows from the fundamental theory of calculus.

Lemma 7.8. For $f \in \Omega^{k}$, we have the Ansatz

$$
\left(W_{k} f\right)(x)=\frac{1}{|G|} \sum_{w \in G} \int_{0}^{1} q_{w}(t)\langle w x, f(t w x)\rangle d t .
$$

If $\left(\nabla_{k} W_{k} f\right)(x)=f(x)$, then the functions $q_{w}(t)$ are the unique solution of linear system of differential equation given by

$$
\begin{equation*}
t q_{w}^{\prime}(t) w f(t w x)=\sum_{\alpha \in R_{+}} k_{\alpha}\left[q_{w}(t)-q_{r_{\alpha} w}(t)\right] w f(t w x), \tag{7.5}
\end{equation*}
$$

for $w \in G$, with boundary conditions $q_{1}(1)=|G|$ and $q_{w}(1)=0$ for $w \neq 1$.
Proof. To prove this, we want to rewrite $\nabla_{k} W_{k}(f)$ in the form of the left hand side of (7.4), so we can compute the integral. To do this we need to integrate $(7.4)$ in the variable $t$ over the interval $[0,1]$ and set this equal to $(7.2)$. This gives

$$
\left(\nabla_{k} W f\right)(x)=\frac{1}{|G|} \sum_{w \in G} \int_{0}^{1}\left(\frac{d}{d t} t q_{w}(t) w f(w x t)\right) d t
$$

so

$$
\begin{aligned}
& \frac{1}{|G|} \sum_{w \in G}\left(q_{w}(t)+q_{w}(t) t \frac{\partial}{\partial t}+t q_{w}^{\prime}(t)\right) w f(w x t) \\
& \quad=\frac{1}{|G|} \sum_{w \in G} \int_{0}^{1}\left(\left(q_{w}(t)+q_{w}(t) t \frac{\partial}{\partial t}+\sum_{\alpha \in R_{+}} k_{\alpha}\left[q_{w}(t)-q_{r_{\alpha} w}(t)\right]\right) w f(t w x) d t,\right.
\end{aligned}
$$

and the functions $q_{w}(t)$ must satisfy the differential equation

$$
\sum_{w \in G} t q_{w}^{\prime}(t) w f(t w x)=\sum_{w \in G} \sum_{\alpha \in R_{+}} k_{\alpha}\left[q_{w}(t)-q_{r_{\alpha} w}(t)\right] w f(t w x) .
$$

Since this must be valid for all $f \in \Omega^{k}$, the $m$-tuples $f(t w x)$ are linear independent and this gives the system in the Lemma.
From (7.4) follows that

$$
\left(\nabla_{k} W_{k} f\right)(x) \frac{1}{|G|} \int_{0}^{1} \sum_{w \in G}\left(\frac{d}{d t} t q_{w}(t) w f(w x t)\right) d t=f
$$

if and only if $q_{1}(1)=|G|$ and $q_{w}(1)=0$ for $w \neq 1$, which gives the boundary condition.
Lemma 7.9. The unique solution $q_{w}(t), w \in G, 0<t \leq 1$ of (7.5), with boundary conditions $q_{1}(1)=|G|$ and $q_{w}(1)=0$ for $w \neq 1$, is given by

$$
\sum_{w \in G} q_{w}(t) w=|G| \exp \left((\log t) \sum_{\alpha \in R_{+}} k_{\alpha}\left(1-r_{\alpha}\right)\right) .
$$

Here both sides are viewed as elements of the group algebra $\mathbb{C} G$. The exponent on the right hand side can be computed by the usual power series and this power series converges in $\mathbb{C} G$.

Proof. For solving (7.5), we try a solution of the form

$$
\begin{equation*}
\sum_{w \in G} q_{w}(t) w=c \exp \left(\log (t) \sum_{z \in G} a_{z} z\right), \tag{7.6}
\end{equation*}
$$

with $a_{z} \in \mathbb{R}$ and $c \in \mathbb{R}$, as a solution of (7.5). This gives

$$
\sum_{w \in G} t \frac{d}{d t} q_{w}(t) w=c \exp \left(\log (t) \sum_{z \in G} a_{z} z\right)\left(\sum_{w \in G} a_{w} w\right)
$$

which shows that

$$
t q_{w}^{\prime}(t) w=c \sum_{z \in G} q_{z^{-1} w}(t) a_{z} w
$$

By plugging this into (7.5), we get

$$
\sum_{z \in G} q_{w z^{-1}}(t) a_{z}=\sum_{\alpha \in R_{+}} k_{\alpha}\left(q_{w}(t)-q_{r_{\alpha} w}(t)\right),
$$

so $a_{1}=\gamma_{k}, a_{r_{\alpha}}=k_{\alpha}$ and $a_{z}=0$ for all other $z \in G$, where we have used that $r_{\alpha} w=w r_{\beta}$ for some conjugate root $\beta \in R_{+}$, so $k_{\alpha}=k_{\beta}$. To compute the constant $c$, note that

$$
\sum_{w \in G} q_{w}(1) w=c \exp \left(\log (1) \sum_{z \in G} a_{z} z\right)=c \cdot 1(\in \mathbb{C} G),
$$

so 7.6 solves 7.5 if $c=|G|$. Next note that $\lim _{t \rightarrow 0} \exp \left((\log t) \sum_{\alpha \in R_{+}} k_{\alpha}\left(1-r_{\alpha}\right)\right)=0$, so $q_{w}(t)$ is continuous at 0 .

Note that the functions $q_{w}(t)$ in Lemma 7.9 are equal to the functions $p_{w}(t)$ defined in Definition 6.45.

### 7.2 The construction of V

In the previous section we have found the operator $W_{k}: \Omega \rightarrow C^{2}\left(\mathbb{R}^{m}\right)$, which is the inverse of $\nabla_{k}$. We can use this operator to compute the intertwining operator $V$. We will also have a look at some properties of the functions $p_{w}(t):(0,1] \rightarrow \mathbb{R}$.

Lemma 7.10. [10, Lemma 3.5] The functions $p_{w}(t), w \in G$ satisfy the following:

$$
\begin{equation*}
t p_{w}^{\prime}(t)=\sum_{\alpha \in R_{+}} k_{\alpha}\left(p_{w}(t)-p_{w r_{\alpha}}(t)\right) \tag{7.7}
\end{equation*}
$$

$p_{1}(1)=G, p_{w}(1)=0$ for $w \neq 1$, and $\sum_{w \in G} p_{w}(t)=|G|, 0<t \leq 1$. Also $p_{w}(t) \geq 0$ for $0<t \leq 1$.

Proof. By applying $t \frac{d}{d t}$ to the formula in Definition 6.45, we find

$$
\begin{aligned}
t \frac{d}{d t} \sum_{w \in G} p_{w}(t) w & =t|G| \exp \left((\log t) \sum_{\alpha \in R_{+}} k_{\alpha}\left(1-r_{\alpha}\right)\right)\left(\sum_{\alpha \in R_{+}} k_{\alpha}\left(1-r_{\alpha}\right)\right) / t \\
& =t \sum_{\alpha \in R_{+}} k_{\alpha} \sum_{w \in G} p_{w}(t) t^{-1}\left(w-w r_{\alpha}\right) \\
& =\sum_{w \in G} \sum_{\alpha \in R_{+}} k_{\alpha}\left(p_{w}(t)-p_{w r_{\alpha}}(t)\right) w
\end{aligned}
$$

which implies (7.7).
Next, note that $1 /|G| \sum_{w \in G} p_{w}(1) w=\exp (0)=1 \in \mathbb{C} G$, so $p_{1}(1)=1$ and $p_{w}(1)=0$ if $w \neq 1$. Let $\chi$ be the trivial character 1. Then

$$
\psi_{1}\left(\sum_{w \in G} p_{w}(t) w\right)=\sum_{w \in G} p_{w}(t)=|G| \exp \left((\log t) \sum_{\alpha \in R_{+}} k_{\alpha}(1-1)\right)=|G| .
$$

Finally, we find from Definition 6.45 that

$$
\sum_{w \in G} p_{w}(t) w=|G|\left(\left(t^{\sum_{\alpha \in R_{+}} k_{\alpha}}\right) 1\right)\left(\exp \left((-\log t) \sum_{\alpha \in R_{+}} k_{\alpha} r_{\alpha}\right)\right) .
$$

Since the argument of the exponential function is positive for $0<t \leq 1$, we see that $p_{w}(t) \geq 0$ for each $w \in G$ for $0 \leq t \leq 1$.

Theorem 7.11. [10, Thm. 3.8] Let $f$ be a $k$-exact m-tuple of $C^{1}$-functions on $\left\{x \in \mathbb{R}^{m}\right.$ : $|x|<r\}$ for some $r>0$. Define $F \in C^{2}\left(\mathbb{R}^{m}\right)$ by

$$
F(x)=\frac{1}{|G|} \int_{0}^{1} \sum_{w \in G} p_{w}(t)\langle w x, f(w x t)\rangle d t
$$

then $\nabla_{k} F=f$ for $|x|<r$ and $F(0)=0$.
Proof. By applying $\nabla_{k}$ to $F$, interchanging $\nabla_{k}$ with the integral and using Corollary 7.6, we find

$$
\begin{aligned}
& \nabla_{k} F(x) \\
& =\frac{1}{|G|} \sum_{w \in G} \int_{0}^{1} p_{w}(t)\left[w f(t w x)+w t \frac{\partial}{\partial t} f(t w x)+\sum_{\alpha \in R_{+}} k_{\alpha}\left(w f(t w x)-r_{\alpha} w f\left(t w r_{\alpha} x\right)\right)\right] d t \\
& =\frac{1}{|G|} \sum_{w \in G} \int_{0}^{1} w\left[p_{w}(t) f(t w x)+t \frac{\partial}{\partial t} f(t w x) \sum_{\alpha \in R_{+}} k_{\alpha}\left(p_{w}(t)-p_{w r_{\alpha} x}(t)\right) f(t w x)\right] d t,
\end{aligned}
$$

where we have rewritten the term $\sum_{w} p_{w}(t) r_{\alpha} w^{-1} f\left(t w r_{\alpha} x\right)$ as $\sum_{w} p_{w r_{\alpha}}(t) w^{-1} f(t w x)$ by changing the summation variable.

By Lemma 7.10 the sum over $\alpha \in R_{+}$is equal to $t p_{w}^{\prime}(t) w f(t w x)$, which leads to

$$
\begin{aligned}
\nabla_{k} F(x) & =1 /|G| \sum_{w \in G} \int_{0}^{1} w\left[p_{w}(t)\left(f(t w x)+t \frac{\partial}{\partial t} f(t w x)\right)+t p_{w}^{\prime}(t) f(t w x)\right] d t \\
& =1 /|G| \sum_{w \in G} \int_{0}^{1} \frac{\partial}{\partial t}\left(t p_{w}(t) w f(t x w)\right) d t \\
& =1 /|G| \sum_{w \in G} p_{w}(1) w f(x w)=f(x)
\end{aligned}
$$

where we have used that $p_{1}(1)=|G|$ and $p_{w}(1)=0$ for $w \neq 1$.
Corollary 7.12. [10, Cor. 3.9] Suppose $f$ and $g$ are $k$-exact 1 -forms, such that $\langle x, f(x)\rangle=$ $\langle x, g(x)\rangle$ for $|x|<r$, then $f=g$.
Proof. Applying Theorem 7.11 to the function $f-g$ gives

$$
F(x)=1 /|G| \sum_{w \in G} \int_{0}^{1} p_{w}(t)\langle w x,(f-g)(w x t)\rangle d t=0
$$

Then $(f-g)(x)=\nabla_{k} F=0$, so $f=g$.
Theorem 7.13. [10, Thm. 3.10] Suppose $F$ is a $C^{2}$-function on $\left\{x \in \mathbb{R}^{m}:|x|<r\right\}$, for some $r>0$. Then

$$
F(x)-F(0)=1 /|G| \sum_{w \in G} \int_{0}^{1} p_{w}(t)\left\langle w x,\left(\nabla_{k} F\right)(w x t)\right\rangle d t
$$

for $|x|<r$.
Proof. Evaluating $\sum_{i=1}^{m} x_{i} T_{i} F(x)$ at $t w x$ and dividing by $t$ gives

$$
\sum_{i=1}^{m}(w x)_{i} T_{i} F(t w x)=\frac{\partial}{\partial t} F(t w x)+t^{-1} \sum_{\alpha \in R_{+}} k_{\alpha}\left(F(t w x)-F\left(t w r_{\alpha} x\right)\right)
$$

for $|x|<r$ and $w \in G$. By using this identity in the integral we find

$$
\begin{aligned}
& \sum_{w \in G} \int_{0}^{1} p_{w}(t)\left\langle\left(w x, \nabla_{k} F\right)(t w x)\right\rangle d t \\
& \quad=\sum_{w \in G} \int_{0}^{1} p_{w}(t)\left[\frac{\partial}{\partial t} F(t w x)+t^{-1} \sum_{\alpha \in R_{+}} k_{\alpha}\left(F(t w x)-F\left(t w r_{\alpha} x\right)\right)\right] d t \\
& \quad=\sum_{w \in G} \int_{0}^{1}\left[p_{w}(t) \frac{\partial}{\partial t} F(t w x)+t^{-1} \sum_{\alpha \in R_{+}} k_{\alpha} F(t w x)\left(p_{w}(t)-p_{w r_{\alpha}}(t)\right)\right] d t \\
& \quad=\sum_{w \in G}\left(p_{w}(1) F(x w)-p_{w}(0) F(0)\right) \\
& \quad=|G|(F(x)-F(0))
\end{aligned}
$$

where (7.7) and the properties $p_{1}(1)=|G|, p_{w}(1)=0$, for $w \neq 1$ and $\sum_{w \in G} p_{w}(t)=|G|$, which were stated in Lemma 7.10 .

Since we have finally found the correct generalization of Lemma 7.3 we can define the intertwining operator $V$ in an inductive way on $P$.

Theorem 7.14. [10, Thm. 3.11] There exists a unique linear map $V: P \rightarrow P, V(1)=1$, $T_{i} V f(x)=V\left(\partial_{i} f\right)(x)$ and $V P_{n} \subset P_{n}$. The map $V$ is invertible.
Proof. We will define operators $V_{n}: P_{n} \rightarrow P_{n}$ by recursion over $n$.
First $V_{0}=\left.I\right|_{P_{0}}$. Assume $V_{n}: P_{n} \rightarrow P_{n}$ has been defined, then define $V_{n+1}: P_{n+1} \rightarrow P_{n+1}$ by

$$
\begin{equation*}
V_{n+1} f(x)=\frac{1}{|G|} \sum_{w \in G} \int_{0}^{1} p_{w}(t) \sum_{i=1}^{m}(w x)_{i}\left(V_{n} \frac{\partial}{\partial x_{i}} f\right)(t w x) d t, \tag{7.8}
\end{equation*}
$$

for $f \in P_{n+1}$. We can rewrite (7.8) as

$$
V_{n+1} f(x)=\frac{1}{|G|} \sum_{w \in G} \sum_{i=1}^{m}(w x)_{i}\left(V_{n} \frac{\partial}{\partial x_{i}} f\right)(w x) \int_{0}^{1} t^{n} p_{w}(t) d t,
$$

and from this it can be seen that $V_{n+1} f \in P_{n+1}$, since each integral gives a constant and all terms in the sum are polynomials of degree $n+1$.
By induction we will show that $T_{i} \circ V_{n}=V_{n-1} \partial_{i}$ on $P_{n}$. This statement is true for $n=0$, if we put $V_{-1}=0$.
Assume the statement has been established for $n \geq 0$. Let $p \in P_{n+1}$. Then $V_{n}\left(\partial_{i} f\right)$ is $k$-exact, because $T_{j} V_{n}\left(\partial_{i} f\right)=V_{n-1}\left(\partial_{j} \partial_{i} f\right)=V_{n-1}\left(\partial_{i} \partial_{j} f\right)=T_{i} V_{n}\left(\partial_{j} f\right)$. Since $V_{n}\left(\partial_{i} f\right)$ is $k$-exact, we can apply Theorem 7.11 to the right hand side of (7.8) and conclude that $\nabla_{k} V(f)=V \nabla f$, so $T_{i} V(f)=V \partial_{i} f$.
Next we will prove the uniqueness of $V$ by induction on $n$.
Suppose both $V$ and $V^{\prime}$ have the properties mentioned in Theorem (7.14). Denote $\left.V\right|_{P_{n}}$ by $V_{n}$ and $\left.V^{\prime}\right|_{P_{n}}$ by $V_{n}^{\prime}$.
For $n=0$, we have that $V_{0}(1)=1=V_{0}^{\prime}(1)$, so $V_{0}=V_{0}^{\prime}$.
Let $n>1$. Assume $V_{n}=V_{n}^{\prime}$. Since $T_{i} V_{n+1} f(x)=V_{n}\left(\partial_{i} f\right)(x)$ and $T_{i} V_{n+1}^{\prime} f(x)=V_{n}\left(\partial_{i} f\right)(x)$ for all $f \in P_{n+1}$, we have that

$$
T_{i}\left(V_{n+1}-V_{n+1}^{\prime}\right) f(x)=\left(V_{n}-V_{n}\right)\left(\partial_{i} f\right)(x)=0,
$$

for all $f \in P_{n+1}$ and all $1 \leq i \leq m$. So $\left(V_{n+1}-V_{n+1}^{\prime}\right) f(x) \in \cap_{i=1}^{m} \operatorname{ker} T_{i}=P_{0}$, because $k$ is nondegenerate. Because $V_{n+1}$ and $V_{n+1}^{\prime}$ is homogeneous of degree 0 , this means that $f \in P_{0}$, but $P_{0} \cap P_{n+1}=0$, so $f=0$ and $V_{n+1}=V_{n+1}^{\prime}$.
So because $V_{0}$ is unique, we find by induction that $V$ is unique.
To show that $V$ is invertible, we need to show that $V$ is bijective. We will use induction over $n$. First note that $V_{0}$ is the identity map, so it is bijective.
Next let $n>1$ and assume that $V_{n}$ and $V_{n-1}$ is invertible. We denote by $V_{-1}$ the restriction of $V$ to 0 , given by $V(0)=0$.
For $f \in P_{n+1}$, the $m$-tuple $T_{i} f$ is $k$-exact. Since $V_{n}$ is invertible, we can write $T_{i} f=V_{n} g_{i}$, for a unique $g_{i} \in P_{n}$. At least one of the $g_{i}$ is nonzero, because $k$ is nondegenerate. Then $V_{n-1}^{-1} T_{j} T_{i} f=T_{j} V_{n} g_{i}=V_{n-1}^{-1} V_{n-1} \partial_{i} d_{j}=\partial_{j} g_{i}$, because $T_{i}$ and $T_{j}$ commute. So $g_{i}$ is an exact $m$-tuple. This means that there is some $g \in P_{n+1}$, such that $\partial_{i} g=g_{i}$ and $V_{n+1} g=f$, so $V$ is surjective.
The map $V_{n+1}: P_{n+1} \rightarrow P_{n+1}$ is a surjective linear map. Since $\operatorname{dim}\left(\operatorname{ker}\left(V_{n+1}\right)=\operatorname{dim}\left(P_{n+1}\right)-\right.$ $\operatorname{dim}\left(P_{n+1}\right)=0$, we see that $V_{n+1}$ is injective.
So $V_{n+1}$ is bijective and $V$ is invertible.

Finally we will mention some results of Dunkl (see [11), which show that the intertwining operator $V$, given by $(7.8)$ has an extension to the space $A$, which is a subset of the space of formal power series. We shall define $A$ in Definition 7.17 .

Definition 7.15. [11, Def. 2.4] We define the following three norms $\|\cdot\|_{\infty},\|\cdot\|_{T}$ and $\|\cdot\|_{\partial}$ on the space $P_{n}=P_{n}\left(\mathbb{R}^{m}\right)$. Let $\|c\|_{T}=\|c\|_{\partial}=|c|$, for $c \in P_{0}$. For $p \in P_{n}$, let

$$
\begin{aligned}
& \|p\|_{T}=\frac{1}{n!} \sup _{\left|u_{1}\right|=\cdots=\left|u_{n}\right|=1}\left\|\left(\prod_{i=1}^{n} T_{u_{i}}\right) p\right\|_{T}, \\
& \|p\|_{\partial}=\frac{1}{n!} \sup _{\left|u_{1}\right|=\cdots=\left|u_{n}\right|=1}\left\|\left(\prod_{i=1}^{n} \partial_{u_{i}}\right) p\right\|_{\partial},
\end{aligned}
$$

and

$$
\|p\|_{\infty}=\sup _{|x| \leq 1}|p(x)| .
$$

In [11] is shown that $\|f\|_{\delta}=\|f\|_{\infty} \leq\|f\|_{T}$, for $f \in P_{n}$.
Lemma 7.16. [11, Prop. 2.5] For $p \in P_{n},\|V p\|_{T}=\|p\|_{\delta}$,
Proof. By repeated use of the rule $V \partial_{i}=T_{i} \partial V$, we find for all $u_{i} \in S, 1 \leq i \leq n$ that

$$
\left\|\left(\prod_{i=1}^{n} T_{u_{i}}\right) V p\right\|_{T}=\left\|V\left(\prod_{i=1}^{n} \partial_{u_{i}}\right) p\right\|_{\partial}
$$

The supremum over $u_{i} \in S, 1 \leq i \leq n$ of the left hand side is equal to the supremum over $u_{i} \in S, 1 \leq i \leq n$ of the right hand side, because both sides are elements of $\mathbb{R}$, for fixed $u_{i} \in S, 1 \leq i \leq n$.

Definition 7.17. [11, p. 1217] Let $f=\sum_{n=0}^{\infty} f_{n}$, be a formal sum with $f_{n} \in P_{n}$. Define the norm of the formal sum by

$$
\|f\|_{A}=\sum_{n=0}^{\infty}\left\|f_{n}\right\|_{\infty}
$$

Let $A$ be the space

$$
\left\{f=\sum_{n=0}^{\infty} f_{n}: f_{n} \in P_{n},\|f\|_{A}<\infty\right\} .
$$

Theorem 7.18. [11, Thm 2.6] The operator $V$ extends to a bounded operator on $A$, where $V f=\sum_{n=0}^{\infty} V f_{n}$, for $f=\sum_{n=0}^{\infty} f_{n} \in A,\|V f\|_{A} \leq\|f\|_{A}$ and $|V f(x)| \leq \sum_{n=0}^{\infty}|x|^{n}\left\|f_{n}\right\|_{\infty} \leq$ $\|f\|_{A}, \quad(|x| \leq 1)$.

See [11] for a proof of this theorem.

### 7.3 Degenerate values of $k_{\alpha}$

In this section we will shortly look at the possible values of $k_{\alpha}$, such that $k_{\alpha}$ is nondegenerate. We will consider the Dunkl operators restricted to the space of polynomials in $m$-variables.

Definition 7.19. For a root system $R$, the weight function $k_{\alpha}$ is degenerate if

$$
\cap_{u \in \mathbb{R}^{m}} \operatorname{ker} T_{u} \supsetneq P_{0} .
$$

Lemma 7.20. Let $R$ be a root, with weight function $k_{\alpha}$, and associated Dunkl operators $T_{i}$ and associated intertwining operator $V$. Then $k$ is degenerate if and only if $V$ is not injective.

Proof. Let $k_{\alpha}$ be degenerate. Then $\cap_{u \in \mathbb{R}^{m}}$ ker $T_{u} \supsetneq P_{0}$, so there is a polynomial $p$ of degree at least one, such that $T_{u}(p)=0$, for all $u \in \mathbb{R}^{m}$. Then $V T_{u}(p)=0, \forall u$, so $\partial_{u} V(p)=0, \forall u$, which shows that $V(p)=c \in P(0)$. This shows that $V$ is not injective, because $V(p(x))=V(c)$ and $p(x) \neq c$.
For the converse, assume that $V$ is not injective. Let $N_{0}$ be the set of $n \in \mathbb{N}$, such that $V_{n}=V_{P_{n}}$ is not injective. Note that $0 \notin N_{0}$, because $V(1)=1$ and $V$ is linear.
Let $n$ be smallest element of $N_{0}$, then there are some polynomials $p, q \in P_{n}$, such that $V(p)=V(q)$ and $p \neq q$.
Then $\partial_{u} V(p-q)=0$, so $V T_{u}(p-q)=0$. However $V_{n-1}$ is injective, because $n$ is the smallest element of $N_{0}$. So $V T_{u}(p-q)=$ implies that $T_{u}(p-q)=0$, and $p-q \in\left(\cap_{u \in \mathbb{R}^{m}} \operatorname{ker} T_{u}\right) \backslash P_{0}$, so $k$ is degenerate.

For an example of a degenerate weight function, we can look back at Example 6.7. For the root system $A_{1}$, we find $T_{e_{1}} x^{2 n+1}=(2 n+1+2 k) x^{2 n}$, which shows that $k$ is degenerate if $k_{\sqrt{2}}=-i-1 / 2, i \in \mathbb{N}$.
The inverse of the intertwining operator can be defined by $V^{-1} p=W V^{-1} \nabla_{k} p$, which gives zero for $F=x^{2 i+1}$, which means that $V q \notin P_{2 i+1}$ for some $q \in P_{2 i+1}$.
To get some more information we can have another look at the coefficients $p_{w}(t)$. Recall from Equation (6.13) that

$$
p_{w}(t)=\sum_{\chi \in \widehat{G}} \chi(1) \bar{\chi}(w) t^{\lambda(\chi)}
$$

Inserting this in the formula for the intertwining operator gives

$$
\begin{aligned}
V_{n+1} f(x) & =\frac{1}{|G|} \sum_{\chi \in \widehat{G}} \sum_{w \in G} \sum_{i=1}^{m}(w x)_{i}\left(V_{n} \frac{\partial}{\partial x_{i}} f\right)(w x) \int_{0}^{1} t^{n} \chi(1) \bar{\chi}(w) t^{\lambda(\chi)} d t \\
& =\frac{1}{|G|} \sum_{w \in G}\left\langle w x, V_{n} \nabla f(w x)\right\rangle \sum_{\chi \in \widehat{G}} \frac{\chi(1) \bar{\chi}(w)}{\lambda(\chi)+n+1}
\end{aligned}
$$

for a polynomial $f$ of degree $n$, so we see that the construction of the intertwining operator might break down if for some character $\chi$, the eigenvalue $\lambda(\chi)$ is a negative integer. As an example consider the one dimensional representation $\rho$ defined by $\rho(1)=1$ and $\rho\left(r_{\alpha}\right)=-1$, with character $\chi_{\rho}$. By equation (6.11)

$$
\lambda\left(\chi_{\rho}\right)=2 \sum_{\alpha \in R_{+}} k_{\alpha}=2 \gamma_{k},
$$

so we see that the construction of the intertwining operator might break down if $2 \gamma_{k}=$ $-1,-2, \ldots$.

Lemma 7.21. [12, p. 126] Let $R$ be a rootsystem and let $k$ be a weight function on $R$. Recall from Lemma 6.15 on page 34, that

$$
\Delta_{k}|x|^{2 l} h_{n}=c_{k, n l}|x|^{2(l-1)} h_{n},
$$

with $c_{k, n l}=4(l)\left(n+m_{k} / 2+l-1\right)$, for $l \neq 0 \in \mathbb{N}$ and $n \in \mathbb{N}$. If one of these constants equals 0 , the weight function $k$ is degenerate.

Proof. First assume that for the weight functions $k$, all the constants $c_{k, n l}$ are nonzero and $V$ is injective. Define $A_{k, n j}=\oplus_{i=0}^{j-1}|x|^{2 i} H_{k, n-2 i}$ and $B_{k, n j}=\oplus_{i=j}^{\lfloor n / 2\rfloor}|x|^{2 i} H_{k, n-2 i}$. Note that $P_{n}=A_{k, n j} \oplus B_{k, n j}$, for all $0 \leq j \leq\lfloor n / 2\rfloor$.
We also define the harmonic analogues $A_{n j}=\oplus_{i=0}^{j-1}|x|{ }^{2 i} H_{n-2 i}$ and $B_{n j}=\oplus_{i=j}^{\lfloor n / 2\rfloor}|x|^{2 i} H_{n-2 i}$. Since $\nabla_{k}^{j} A_{k, n j}=0$, we have that $V\left(A_{k, n j}\right) \subset A_{n j}$. By the injectivity of V , we find that $V\left(A_{k, n j}\right) \subset A_{n j}$.
Next assume that for the weight functions $k$, the constant $c_{k, n_{0} l_{0}}=0$ for some $l_{0}$ and $n_{0}$. Then $c_{k, n 1}=0$, for $n=n_{0}+l_{0}-1$. From this it follows that $\nabla_{k} A_{n 1}=0$, so $V\left(A_{n 1} \subset A_{n 0}\right.$. We assumed that $V$ was injective, so we must have that $V\left(A_{n 1}\right)=A_{n 1}$. This leads to contradiction, because $A_{n 1}$ isn't a subset of $A_{n 0}$. It follows that $V$ cannot be injective and by Lemma 7.20 it follows that $k$ is degenerate if $c_{k, n l}=0$ for some $n, l$.

From the previous lemma it follows that $k$ is degerenerate if $c_{k, n l}=4 l\left(n+\gamma_{k}+m / 2+l-1\right)=0$, for some $l \geq 1, n \geq 0 \in \mathbb{N}$. from (6.5). Tt follows that $m_{k} / 2=\gamma_{k}+m / 2 \notin-\mathbb{N}$.
However, finding the specific set of degenerate values requires, more information about the characters of $G$ and a complete treatment of this problem was given in [13.

## Chapter 8

## The Fischer decomposition with respect to $p(T)$ and $p(x)$

In this chapter we want to prove that the Fischer decomposition with respect to $p(T)$ and $p(x)$ exist. Here $p(x)$ is an arbitrary homogeneous polynomial in the variables $x_{1}, \ldots, x_{m}$ and $p(T)$ is the operator formed by replacing $x_{i}$ with $T_{i}$ in $p(x)$. We will prove the existence of this Fischer decomposition by constructing an appropriate inner product and applying Theorem 4.8. To do this we need an inner product similar to the one used in Chapter 5. which can be found using the intertwining operator $V$ defined in equation (7.8).

Definition 8.1. Let $q$ be a polynomial in $m$ variables, $x_{1}, \ldots, x_{m}$. Define by $q(T)$ the power series, which is obtained by replacing the variable $x_{i}$ by the Dunkl operator $T_{i}$. When necessary, we will use the notation $q\left(T_{x}\right)$ to emphasize that we use the Dunkl operators in the variables $x_{1}, \ldots, x_{m}$.

Since the Dunkl operators commute, we can view the polynomial $q(T)$ as an element of the polynomial ring $R=P\left(T_{1}, \ldots, T_{m}\right)$. We have the natural action of $R$ on $P=P\left(\mathbb{R}^{m}\right)$, given by

$$
(r, q) \rightarrow r\left(T_{x}\right) q(x) .
$$

By using multi-index notation we can write

$$
r=\sum_{n=0}^{l} \sum_{|\alpha|=n} c_{\alpha} T^{\alpha}\left(c_{\alpha} \in \mathbb{R}\right)
$$

where $l$ is the degree of $r$.

Definition 8.2. [11, Def 3.1] For $x, y \in \mathbb{R}^{m}$, define $K(x, y)=V_{x}(\exp (\langle x, y\rangle))$. Here the subscript $x$ indicates the variable with respect to which the operator is applied. The operator $V$ was defined in (7.8). Also define $K_{n}(x, y)=V_{x}\left(\langle x, y\rangle^{n} / n!\right)$.

Lemma 8.3. [11, Prop 3.2] For $n \in \mathbb{N}, x, y \in \mathbb{R}^{m}$, some useful properties of the kernel are given by

$$
\text { 1. } K_{n+1}(x, y)=\frac{1}{|G|} \sum_{w \in G}\langle w x, y\rangle K(w x, y) \int_{0}^{1} p_{w}(t) t^{n} d t \text {, for } n \geq 1 \text { and } K_{0}(x, y)=1 \text {, }
$$

2. $\left|K_{n}(x, y)\right| \leq \max _{w \in G}|\langle w x, y\rangle|^{n} / n!$,
3. $K_{n}(w x, w y)=K_{n}(x, y) ; w \in G$,
4. $K_{n}(x, y)=K_{n}(y, x)$,
5. $\left(T_{i}\right)_{x} K_{n}(x, y)=K_{n-1}(x, y) y_{i}$.

Proof. We shall use the relation in part 1, to prove the other results.
Take $f(x)=\langle x, y\rangle^{n+1} /(n+1)!$. Then $\partial_{i} f=y_{i}\langle x, y\rangle^{n} /(n)$ !, so $V_{x}\left(\partial_{i} f\right)=y_{i} K_{n}(x, y)$. Now we can use (7.8) and we see that

$$
\begin{aligned}
V_{x}(f(x)) & =\frac{1}{|G|} \sum_{w \in G} \int_{0}^{1} p_{w}(t) \sum_{i=1}^{m}(w x)_{i}\left(V_{x} \partial_{i} f\right)(w x t) d t \\
& =\frac{1}{|G|} \sum_{w \in G} \int_{0}^{1} p_{w}(t) \sum_{i=1}^{m}(w x)_{i} y_{i} K_{n}(w x, y) t^{n} d t \\
& =\frac{1}{|G|} \sum_{w \in G}\langle w x, y\rangle K_{n}(w x, y) \int_{0}^{1} p_{w}(t) t^{n} d t .
\end{aligned}
$$

For part 2, we will use induction. The estimate is clearly true for $K_{0}(x, y)$. Assume that $\left|K_{n}(x, y)\right| \leq \max _{w \in G}|\langle x, y\rangle|^{n} / n$ ! for some $n \in \mathbb{N}$, then $\left|K_{n}(w x, y)\right| \leq \max _{w \in G}\langle x, y\rangle^{n} / n$ !. We can use part 1 to write

$$
\begin{aligned}
\left|K_{n+1}(x, y)\right| & \leq \frac{1}{|G|} \sum_{w \in G}|\langle w x, y\rangle||K(w x, y)| \int_{0}^{1}\left|p_{w}(t)\right| t^{n} d t \\
& =\left(\max _{w \in G}|\langle w x, y\rangle|^{n+1} / n!\right)\left(\sum_{w \in G} \int_{0}^{1}\left|p_{w}(t)\right| t^{n} d t\right), \\
& =\max _{w \in G}|\langle x, y\rangle|^{n}+1 /(n+1)!,
\end{aligned}
$$

where was used that the functions $p_{w}(t)>0$, for $0 \leq t \leq 1$ and $\sum_{w \in G} p_{w}(t)=G$ (see Lemma 7.10).

For part 3,

$$
\begin{aligned}
K_{n}(w x, w y) & =L\left(w^{-1}\right) K_{n}(x, w y)=V_{x} L\left(w^{-1}\right)\langle x, w y\rangle^{n} / n! \\
& =V_{x}\langle w x, w y\rangle^{n} / n!=V_{x}\langle x, y\rangle^{n} / n!=K_{n}(x, y) .
\end{aligned}
$$

For part 4, we apply induction with respect to $n$. The identity is clear for $n=0$. Assume $K_{n}(x, y)=K_{n}(y, x)$.

$$
\begin{aligned}
K_{n+1}(y, x) & =\frac{1}{|G|} \sum_{w \in G}\langle w y, x\rangle K_{n}(w y, x) \int_{0}^{1} p_{w}(t) t^{n} d t, \\
& =\frac{1}{|G|} \sum_{w \in G}\langle x, w y\rangle K_{n}(x, w y) \int_{0}^{1} p_{w}(t) t^{n} d t, \\
& =\frac{1}{|G|} \sum_{w \in G}\left\langle w^{-1} x, y\right\rangle K_{n}\left(w^{-1} x, y\right) \int_{0}^{1} p_{w}(t) t^{n} d t, \\
& =\frac{1}{|G|} \sum_{w \in G}\langle w x, y\rangle K_{n}(w x, y) \int_{0}^{1} p_{w}(t) t^{n} d t, \\
& =K_{n+1}(x, y) .
\end{aligned}
$$

In the first step we have used the symmetry of $K_{n}$ and the symmetry of the inner product. Next we have used the invariance of the inner product and $K_{n}$ under the Weyl group. Finally we have changed the summation variable from $w$ to $w^{-1}$ and we have used that $p_{w^{-1}}(t)=$ $p_{w}(t)$, since $w$ and $w^{-1}$ are conjugate. (See [10, p.112]) For part 5,

$$
\begin{aligned}
\left(T_{i}\right)_{x} K_{n}(x, y) & =\left(T_{i}\right)_{x} V_{x}\langle x, y\rangle^{n} / n!=V_{x}\left(\partial_{i}\right)_{x}\langle x, y\rangle^{n} / n! \\
& =V_{x} y_{i}\langle x, y\rangle^{n-1} /(n-1)!=y_{i} K_{n-1}(x, y) .
\end{aligned}
$$

Corollary 8.4. [12, p. 127] We can estimate the norm of $K(x, y)=\sum_{n=0}^{\infty} K_{n}(x, y)$ by

$$
|K(x, y)| \leq e^{|x||y|}
$$

for $k_{\alpha}>0$. If $k_{\alpha}$ is nondegenerate but negative for some $\alpha \in R$, instead we have the estimate

$$
|K(x, y)| \leq e^{B|x||y|}
$$

for some $B>0$, depending on $|G|$ and $k_{\alpha}$.
Proof. If $k_{\alpha}>0$, we can use the estimate $|\langle w x, y\rangle| \leq|x||y|$, to write part 2 of Lemma 8.3 as $\left|K_{n}(x, y)\right| \leq|x|^{n}\left|y^{n}\right| / n$ !. By using the Taylor series of the exponent, this leads to

$$
K(x, y) \leq e^{|x|^{n}|y|^{n}} .
$$

The other part is show in [12, p.127].
Corollary 8.5. [11, Cor 3.3] For $p \in P_{n}$, we have that $K_{n}\left(x, T_{y}\right) p(y)=p(x)$.
Proof. Recall from Definition 5.4 that $\hat{K}_{n}(x, y)=\langle x, y\rangle^{n}$.
Let $q \in P_{n}$. By Lemma 5.5, we have $q(x)=K_{n}(x, y) q(y)$. By applying $V_{x}$ to both sides, it follows that

$$
V_{x} q(x)=V_{x} K_{n}\left(x, \partial_{y}\right) q(y) .
$$

By applying $V_{y}$ to both sides and noting that the left hand side is constant in $y$, we find

$$
V_{x} q(x)=V_{x} K_{n}\left(x, T_{y}\right) V_{y} q(y),
$$

because $T_{y} V_{y}=V_{y} \partial_{y}$. So we have proven the corollary for each polynomial of the form $p(x)=V_{x} q(x)$. Because $V: P_{n} \rightarrow P_{n}$ is bijective, we have proven the corollary for each $p \in P_{n}$.

Theorem 8.6. [11, Thm. 3.5] Let $R$ be a root system on $\mathbb{R}^{m}$, with weight function $k$. We assume that $k$ is nondegenerate. Then the form $[\cdot, \cdot]_{k}: P\left(\mathbb{R}^{m}\right) \times P\left(\mathbb{R}^{m}\right) \rightarrow \mathbb{R}$ defined by

$$
[p, q]_{k}=\left.(p(T) q)(x)\right|_{x=0}
$$

defines an inner product on $P\left(\mathbb{R}^{n}\right)$.
It has the useful property that $\left[x_{i} p, q\right]_{k}=\left[p, T_{i} q\right]_{k}$.
Proof. Choose $p, q \in P\left(\mathbb{R}^{n}\right)$ arbitrary. The form is clearly bilinear, so we can assume that $p$ and $q$ are homogeneous polynomials.
We can see that $[p, q]_{k}=0$, if $p$ and $q$ are homogeneous polynomials and $\operatorname{deg} p \neq \operatorname{deg}(q)$. We can also see that $p\left(T_{x}\right) q(x)$, is a scalar if $p$ and $q$ are homogeneous of the same degree, so we can write $\left.[p, q]_{k}=p(T) q\right)(x)$, where we identify $P(0)$ with $\mathbb{R}$.
We need to prove the symmetry of the form. If $p$ and $q$ are homogeneous of different degree, we have $[p, q]_{k}=[q, p]_{k}=0$, so in this case the form is symmetric.
Let $p, q$ be homogeneous polynomials of degree $n$.
By using the generating kernel $K(x, y)$, we can see that

$$
[p, q]_{k}=p\left(T_{x}\right) q(x)=K_{n}\left(T_{x}, T_{y}\right) p(y) q(x)=K_{n}\left(T_{y}, T_{x}\right) q(x) p(y) .
$$

because $T_{x}$ and $T_{y}$ commute and because $K_{n}$ is symmetric, which was shown in Lemma 8.3 part 4. On the other hand

$$
[q, p]_{k}=q\left(T_{y}\right) p(y)=K_{n}\left(T_{y}, T_{x}\right) q(x) p(y)
$$

so we have that $[p, q]_{k}=[q, p]_{k}$, if $p, q$ are homogeneous of degree $n$.
As last step in the proof, we need to prove the positivity of the form. For this we need another lemma.

Lemma 8.7. [11, Thm 3.6] Let $p, q \in P_{n}$ and decompose them as

$$
\begin{aligned}
& p=\sum_{j \leq n / 2}|x|^{2 j} p_{n-2 j}, \\
& q=\sum_{j \leq n / 2}|x|^{2 j} q_{n-2 j},
\end{aligned}
$$

with $p_{n-2 j}, q_{n-2 j} \in H_{k, n-2 j}$.
Then

$$
[p, q]_{k}=\sum_{j \leq n / 2} 4^{j} j!(n-2 j-\gamma+m / 2)_{j}\left[p_{n-2 j}, q_{n-2 j}\right]
$$

Here we have used the notation $(k)_{j}=\prod_{i=0}^{j-1}(k+i)$, where $(k)_{j}$ is a so-called Pochhammer symbol.

Proof. The existence of these decompositions of $p$ and $q$ is given in Theorem 6.16. Using $\Delta_{k}=\sum_{i=1}^{m} T_{i}^{2}$, we find

$$
[p, q]_{k}=\left.\sum_{j \leq n / 2} \sum_{l \leq n / 2} \Delta_{k}^{l} p_{n-2 l}(T)\left(|x|^{2 j} q_{n-2 j}(x)\right)\right|_{x=0} .
$$

Using the commutation relations given in (6.5), we see that

$$
\Delta_{k}|x|^{2 j} f_{n}(x)=4 j(n+j+\gamma-1+m / 2)|x|^{2 j-2} f_{n}(x)+|x|^{2 j} \Delta_{k} f_{n}(x)
$$

and be repeated use of this formula we find

$$
\Delta_{k}^{l}|x|^{2 j} q_{n-2 j}(x)=4^{l}(-j)_{l}(-n+j-\gamma+1-m / 2)_{l}|x|^{2 j-2 l} q_{n-2 j}(x) .
$$

This expression is zero for $l>j$, since it would equal $\Delta^{l-j} C q_{n-2 j}=0$, where $C$ is a some constant. By the same argument this expression is also zero for $l<j$, because the form is symmetric.
So we are left with the $l=j$ terms and this gives

$$
[p, q]_{k}=\left.\sum_{j \leq n / 2} 4^{j} j!(n-2 j+\gamma-m / 2)_{j} p_{n-2 j}(T) q_{n-2 j}(x)\right|_{x=0}
$$

Now we can continue with the proof of Theorem 8.6. We can see that, as long as $\gamma$ is positive, the constants are all positive, which shows that $[p, p]_{k}>0$ for all nonzero polynomials.
Finally the property $\left[x_{i} p, q\right]_{k}=\left[p, T_{i} q\right]_{k}$ follows from the definition in a trivial way.
Theorem 8.8. Let $p(x)$ be an arbitrary homogeneous polynomial of degree l. Define $p(T)$, as the differential difference operator of degree $l$, which is obtained by evaluating $p$ in the point $\left(T_{1}, \ldots, T_{m}\right)$. Define the spaces of ( $\mathrm{p}, \mathrm{k}$ )-harmonics by $H_{k, n}^{p}=P_{n} \cap \operatorname{ker}(p(T))$. Then we have the decomposition

$$
P_{n}=\oplus_{i} p(x)^{i} H_{k, n-l i}^{p} .
$$

Proof. The space $P=\oplus_{n=0}^{\infty} P_{n}$ is a graded vector space and $[\cdot, \cdot]_{k}$ is an inner product on this space and $P_{n} \perp P_{m}$ for $n \neq m$. The operators multiplication by $p(x): P \rightarrow P$ and $p(T): P \rightarrow P$ are formal adjoints with respect to this inner product $[\cdot, \cdot]_{k}$. So the operators satisfy the conditions of Theorem 4.8 and so the space $P_{n}$ can be decomposed as

$$
P_{n}=\oplus_{i} p(x)^{i} H_{k, n-l i}^{p}
$$

This is a also a Fischer decomposition. A special case of this decomposition is given by $p(x)=|x|^{2}$ and $p(T)=\Delta_{k}$.

## Chapter 9

## The Dunkl transform

The Fourier transform is defined by

$$
\hat{f}(x)=\frac{1}{(2 \pi)^{m / 2}} \int_{\mathbb{R}} \hat{K}(-i x, y) f(y) d y=\frac{1}{(2 \pi)^{m / 2}} \int_{\mathbb{R}} \exp (\langle-i x, y\rangle) f(y) d y
$$

for $f \in \mathcal{S}\left(\mathbb{R}^{m}\right)$. Here $\mathcal{S}\left(\mathbb{R}^{m}\right)$ is the Schwartz-space defined by

$$
\mathcal{S}\left(\mathbb{R}^{m}\right)=\left\{f \in C^{\infty}\left(\mathbb{R}^{m}\right):\left.\left|\partial^{\alpha}\right| x\right|^{\beta} f(x) \mid<\infty, \text { for all multi-indices } \alpha, \beta\right\}
$$

We can define a similar transformation by using the kernel $K(x, y)=V_{x}(\exp (\langle x, y\rangle)$. This transformation is called the Dunkl transform. Before we can use it, we first need to determine the domain and range of the transform. We will also use the Laquerre polynomials to construct a set of eigenfunctions of the Dunkl transform. In this chapter we will follow [11] and [12].
Define the function $h: \mathbb{R}^{m} \rightarrow \mathbb{R}$ by

$$
h(x)=\prod_{\alpha \in R_{+}}|\langle\alpha, x\rangle|^{k_{\alpha}}
$$

and define the constants $\gamma=\operatorname{deg}(h)$ and $m_{k}=m+2 \gamma$. These definitions were given before in Definition 6.18.
We also need the measure $h^{2} d \omega$ on the unit sphere $S=\left\{x \in \mathbb{R}^{m}:|x|=1\right\}$, the measure $h^{2} d x$ on $\mathbb{R}^{m}$ and the Gaussian measure

$$
h^{2} d \mu=h^{2}(x)(2 \pi)^{-m / 2} e^{-|x|^{2} / 2} d x
$$

on $\mathbb{R}^{m}$. Here $d \omega$ is the normalized rotation invariant surface measure on the sphere and $d x$ is the Lebesque measure on $\mathbb{R}^{m}$.
We will also use the normalization constants $c_{m}=\left(\int_{\mathbb{R}^{m}} h^{2} d \mu\right)^{-1}$ and $c_{m}^{\prime}=\left(\int_{S^{m-1}} h^{2} d \omega\right)^{-1}$. These measures and normalization constants were defined in Definition 6.19.

Definition 9.1. [12, p. 128] Let

$$
\mathbb{E}\left(\mathbb{R}^{m}\right)=\left\{f \in C^{\infty}\left(\mathbb{R}^{m}\right): \int_{\mathbb{R}^{m}}\left|p\left(\frac{d}{d x_{1}}, \ldots, \frac{d}{d x_{n}}\right) f(x)\right| e^{B|x|} d x<\infty,\right\}
$$

for all $p \in P$ and $B<\infty$.

In Corollary 8.4 we have shown that $K(x, y)<\exp (B|x||y|)$, for some constant $B>0$. In [5. Thm.3.1] it was shown that this estimate holds for complex $x$, so for $f \in \mathbb{E}$ the integral $\int_{\mathbb{R}^{m}} f(x) K(-i x, y)|h(x)|^{2} d x$ exists and so the Dunkl transform is well-defined on $\mathbb{E}$. It was even shown in [22, Prop 2.36] that $|K(-i x, y)| \leq 1$, for $k_{\alpha} \geq 0$, which shows that the transform can be defined for $f \in L^{1}\left(\mathbb{R}^{m}, d x\right)$. However, the set of eigenfunctions of $\mathscr{D}_{k}$ which we are going to construct, consists of elements which are in $\mathbb{E}$ and which are rapidly decreasing at $\infty$, so the proofs do not change. Also note that the functions $p(x) e^{-|x|^{2} / 2}, p \in P$ belong to $\mathbb{E}$.

Definition 9.2. [12, Def. 2.2] For $f \in \mathbb{E}\left(\mathbb{R}^{n}\right)$ and $y \in \mathbb{R}^{n}$ define the Dunkl transform by

$$
\left(\mathscr{D}_{k} f\right)(y)=(2 \pi)^{m / 2} c_{m} \int_{\mathbb{R}^{m}} f(x) K(x,-i y)|h(x)|^{2} d x .
$$

The function $\left(\mathscr{D}_{k} f\right)(y)$ is continuous by the dominated convergence theorem.
Definition 9.3. [23, p.100] The Laquerre polynomials are defined by

$$
L_{n}^{A}(t)=\frac{(A+1)_{n}}{n!} \sum_{j=0}^{n} \frac{(-n)_{j}}{(A+1)_{j}} \frac{t^{j}}{j!},
$$

where we have used the notation $(n)_{j}=n \cdot(n+1) \cdots \cdots(n+j-1)$.
The Laquerre polynomials satisfy the orthogonality relations

$$
\begin{equation*}
\Gamma(A+1)^{-1} \int_{0}^{\infty} L_{k}^{A}(t) L_{l}^{A}(t) t^{A} e^{-t} d t=(A+1)_{k} \delta_{k l} /(n!) \tag{9.1}
\end{equation*}
$$

Definition 9.4. [12, Def. 2.3] For $j, n \in \mathbb{N}$ and $p \in H_{k, n}$ define

$$
\phi_{j}(p ; x)=p(x) L_{j}^{n+\gamma+m / 2-1}\left(|x|^{2}\right) e^{-|x|^{2} / 2}\left(x \in \mathbb{R}^{m}\right) .
$$

We are going to show that the functions $\phi_{j}(p ; x)$ are eigenvectors of the Dunkl transform (see Theorem 9.12). Also, if the polynomials $p_{n, k, i}, 1 \leq i \leq \operatorname{dim}\left(H_{k, n}\right)$ form a basis of $H_{k, n}$, the functions $\phi_{j}\left(p_{n, k, i} ; x\right),(j \in \mathbb{N})$ are orthogonal (see Lemma 9.10 ) and span a dense subset of $L^{2}\left(\mathbb{R}^{m},|h|^{2} d x\right)$ (See Theorem 9.11). To prove all this, we need some calculations.

Lemma 9.5. [11, Thm. 3.8] Let $[\cdot, \cdot]_{k}$ be the inner product defined in Theorem 8.6. Let $p, q \in H_{k, n}$, then

$$
[p, q]_{k}=c_{m} \int_{\mathbb{R}^{m}} p q h^{2} d \mu=c_{m}^{\prime} 2^{n}\left(\frac{N}{2}-\gamma\right)_{n} \int_{S} p q h^{2} d \omega .
$$

Proof. We have that $c_{m} \int_{\mathbb{R}^{m}}|h|^{2} d \mu=1$, by the definition of $c_{m}$.
Since $p\left(T_{x}\right)(q(x))$ is a constant, we can put in inside this integral, which leads to

$$
\begin{aligned}
{[p, q]_{h} } & =c_{m} \int_{\mathbb{R}^{m}}\left(p\left(T_{x}\right) q(x)\right)(1) h^{2} d \mu \\
& =c_{m} \int_{\mathbb{R}^{m}} q(x) p\left(T_{x}\right)^{*}(1) h^{2} d \mu \\
& =c_{m} \int_{\mathbb{R}^{m}} q(x)\left(p(x)+p_{0}(x)\right) h^{2} d \mu
\end{aligned}
$$

Repeated use of $\left(T_{i}\right)^{*} g=x_{i} g(x)-T_{i} g(x)$ and noting that $\operatorname{deg}\left(T_{i} g\right)<\operatorname{deg} g$, shows that the terms of highest degree in $p(T)^{*} 1$ are precisely $p(x)$. By Theorem 6.25 the integral $\int_{\mathbb{R}^{m}} q p_{0} h^{2} d \mu=0$, so we are left with

$$
[p, q]_{k}=c_{m} \int_{\mathbb{R}^{m}} p q h^{2} d \mu=c_{m}^{\prime} 2^{n}\left(\frac{N}{2}-\gamma\right)_{n} \int_{S} p q h^{2} d \omega
$$

where the last equality follows from $(6.6)$.
Lemma 9.6. [11, Prop. 3.9] Let $j, n \in \mathbb{N}$ and $f \in H_{k, n}$, then

$$
\exp \left(-\Delta_{k} / 2\right)|x|^{2 j} f(x)=(-1)^{j} j!2^{j} L_{j}^{m+\gamma+N / 2-1}\left(|x|^{2} / 2\right) f(x)
$$

Here $\exp \left(-\Delta_{k} / 2\right)|x|^{2 j} f(x)=\sum_{i=1}^{\lfloor j+n / 2\rfloor}\left(-\Delta_{k} / 2\right)^{i}|x|^{2 j} f(x)$.
Proof. By Theorem 6.26 it follows that

$$
\begin{aligned}
\exp \left(-\Delta_{k} / 2\right)|x|^{2 j} f(x) & =\sum_{l=0}^{j} \frac{\Delta_{k}^{l}}{(-2)^{l} l!}|x|^{2 j} f(x) \\
& =\sum_{l=0}^{j} \frac{(-2)^{l}}{l!}(-l)_{j}(-m-j-\gamma-N / 2+1)_{j}|x|^{2 j-2 l} f(x) \\
& =\frac{(-1)^{j} j!}{(-1)^{j} j!} \sum_{l=0}^{j} \frac{(-l)_{j}(-j-(m+\gamma+N / 2-1))_{l}}{l!}(-2)^{l} 2^{j-l}\left(|x|^{2} / 2\right)^{j-l} f(x) \\
& =(-1)^{j} j!2^{j} L_{j}^{m+\gamma+N / 2-1}\left(|x|^{2} / 2\right) f(x),
\end{aligned}
$$

where we have used the reversed form of the Langrange polynomials

$$
L_{j}^{A}(t)=\frac{(-1)^{j}}{j!} \sum_{l=0}^{j}(-1)^{l} \frac{(-j)_{l}(-j-A)_{l}}{l!} t^{l-j} .
$$

Theorem 9.7. [11, Thm. 3.10] For $p, q$ polynomials,

$$
[p, q]_{k}=c_{m} \int_{\mathbb{R}^{m}}\left(\exp \left(\Delta_{k} / 2\right) p\right)\left(\exp \left(\Delta_{k} / 2\right) q h^{2} d \mu\right.
$$

Proof. Let $p \in H_{k, a}$ and $q \in H_{k, b}$. Set $A=a+\gamma+m / 2-1$ and $B=b+\gamma+m / 2-1$ and have a look at the integral

$$
I(p, q) \int_{\mathbb{R}^{m}} L_{a}^{A}\left(|x|^{2} / 2\right) L_{b}^{B}\left(|x|^{2} / 2\right) p(x) q(x) h^{2}(x) d \mu, \text { for } j, l \in \mathbb{N} .
$$

Since the polynomials $p, q$ and $h^{2}$ are homogeneous in $|x|$, we can use polar coordinates to get

$$
\begin{aligned}
I(p, q) & =\int_{\mathbb{R}^{m}} L_{j}^{A}\left(|x|^{2} / 2\right) L_{l}^{B}\left(|x|^{2} / 2\right) p(x) q(x) h^{2}(x) d \mu \\
& =\frac{2^{1-m / 2}}{\Gamma(m / 2)} \int_{0}^{\infty}|x|^{a+b+2 \gamma} L_{j}^{A}\left(|x|^{2} / 2\right) L_{l}^{B}\left(|x|^{2} / 2\right) \exp \left(-|x|^{2} / 2\right)|x|^{m-1} d|x| \int_{S} p(x) q(x) h^{2} d \omega .
\end{aligned}
$$

By 6.25 the second integral is 0 if $a \neq b$. Next we use a coordinate transformation $t=|x|^{2} / 2$ to find

$$
\begin{aligned}
I(p, q) & =\int_{\mathbb{R}^{m}} L_{j}^{A}\left(|x|^{2} / 2\right) L_{l}^{A}\left(|x|^{2} / 2\right) \delta_{a b} p(x) q(x) h^{2}(x) d \mu \\
& =\delta_{a b} \frac{2^{1-m / 2}}{\Gamma(m / 2)} \int_{0}^{\infty}|x|^{a+b+2 \gamma} L_{j}^{A}\left(|x|^{2} / 2\right) L_{l}^{A}\left(|x|^{2} / 2\right) \exp \left(-|x|^{2} / 2\right)|x|^{m-1} d|x| \int_{S} p(x) q(x) h^{2} d \omega \\
& =\delta_{a b} \frac{2^{1-m / 2}}{\Gamma(m / 2)} \int_{0}^{\infty}|2|^{a+\gamma+m / 2-1 / 2} t^{a+\gamma+m / 2-1 / 2} L_{j}^{A}(t) L_{l}^{A}(t) \exp (-t) t^{-1 / 2} d t \int_{S} p(x) q(x) h^{2} d \omega \\
& =\delta_{a b} \frac{2^{a+\gamma}}{\Gamma(m / 2)} \int_{0}^{\infty} t^{A} L_{j}^{A}(t) L_{l}^{A}(t) \exp (-t) d t \int_{S} p(x) q(x) h^{2} d \omega \\
& =\delta_{a b} \delta_{j l} \frac{2^{a+\gamma}}{j!} \frac{(A+1)_{j} \Gamma(A+1)}{\Gamma(m / 2)} \int_{S} p(x) q(x) h^{2} d \omega \\
& =\delta_{a b} \delta_{j l} \frac{2^{a+\gamma}}{j!} \frac{\Gamma(a+\gamma+m / 2+k)}{\Gamma(m / 2)} \int_{S} p(x) q(x) h^{2} d \omega .
\end{aligned}
$$

We only need to check the lemma for pairs of monomials $|x|^{2 j} p$ and $|x|^{2 l} q$, with $p \in H_{k, a}$ and $q \in H_{k, b}$. By Lemma 9.6 we can see that

$$
\begin{aligned}
c_{m} & \int_{\mathbb{R}^{m}}\left(\exp \left(\Delta_{k} / 2\right) p\right)\left(\exp \left(\Delta_{k} / 2\right) q\right) h^{2} d \mu \\
& =c_{m}(-2)^{j} j!(-2)^{l} l!I(p, q) \\
& =c_{m}(j!)^{2} \delta_{a b} \delta_{j l} \frac{2^{2 j+a+\gamma}}{j!} \frac{\Gamma(a+\gamma+m / 2+k)}{\Gamma(m / 2)} \int_{S} p(x) q(x) h^{2} d \omega, \\
& =\delta_{a b} \delta_{j l} 2^{a+2 j} \frac{\Gamma(a+\gamma+m / 2+k)}{\Gamma(m / 2+\gamma)} j!c_{m}^{\prime} \int_{S} p(x) q(x) h^{2} d \omega, \\
& =\delta_{a b} \delta_{j l} 4^{j} j!(a+\gamma+m / 2)_{j} 2^{a}(\gamma+m / 2)_{a} c_{m}^{\prime} \int_{S} p(x) q(x) h^{2} d \omega, \\
& =\delta_{j l} 4^{j} j!(a+\gamma+m / 2)_{j}[p, q]_{k}, \\
& =\left[|x|^{2 j} p,|x|^{2 l} q\right]_{k},
\end{aligned}
$$

where we also have used equation (6.6), Theorem 8.6 and equation (9.1).
A simple calculation using Theorem 6.26 shows that

$$
\Delta L_{j}^{n+m / 2+\gamma-1} p(x)=(n+m / 2+\gamma+k-1) L_{j-1}^{n+m / 2+\gamma-1} p(x),
$$

for $j \in \mathbb{N}$ and $p \in H_{k, n}$.
From now on we will follow [12.
For $y \in \mathbb{C}^{m}$ define $\nu(y)=\sum_{i=1}^{m} y_{i}^{2}(\in \mathbb{C})$.
Lemma 9.8. [12, Prop 2.1] Let $p \in P$ and let $y \in \mathbb{C}^{m}$, then

$$
c_{m} \int_{\mathbb{R}^{m}}\left(e^{\Delta_{k} / 2} p(x)\right) K(x, y) h(x)^{2} d \mu(x)=e^{\nu(y) / 2} p(y)
$$

Proof. Let $l$ be an integer larger than the degree of $p$, fix $y \in \mathbb{C}^{m}$ and let

$$
q_{m}(x)=\sum_{j=0}^{l} K_{j}(x, y),
$$

then $\left[q_{l}, p\right]_{k}=p(y)$, which can be seen by breaking $p$ into homogeneous components. This is a polynomial identity which remains valid for a complex $y$. By Theorem 9.7

$$
\left[q_{m}, p\right]_{k}=\int_{\mathbb{R}^{m}}\left(e^{-\Delta_{k} / 2} p\right)\left(e^{-\Delta_{k} / 2} q_{m}\right) h^{2} d \mu
$$

But $\Delta_{k}^{x} K_{n}(x, y)=\nu(y) K_{n-2}(x, y)$ and so

$$
\begin{aligned}
e^{-\Delta_{k} / 2} q_{l}(x) & =\sum_{j=0}^{l} \sum_{r \leq j / 2}\left((-\nu(y) / 2)^{r} / r!\right) K_{j-2 r}(x, y) \\
& =\sum_{r \leq l / 2}\left((-\nu(y) / 2)^{r} / r!\right) \sum_{s=0}^{l-2 r} K_{s}(x, y)
\end{aligned}
$$

By taking the limit $l \rightarrow \infty$, the sum converges to $e^{-\nu(y) / 2} K(x, y)$, since it is dominated termwise by

$$
\sum_{j=0}^{\infty}\left(|y|^{2} / l!2^{l}\right) \sum_{s=0}^{\infty}\left(|x|^{2}|y|^{2} / s!\right)=e^{|y|^{2} / 2+|x||y|}
$$

which is integrable with respect to $d \mu$. So by the dominated convergence theorem

$$
p(y)=e^{-\nu(y) / 2} c_{m} \int_{\mathbb{R}^{m}}\left(e^{-\Delta_{k} / 2} p(x)\right) K(x, y) h^{2}(x) d \mu(x)
$$

Lemma 9.9. [12, Thm. 3.2] For $y, z \in \mathbb{C}^{m}$, we have

$$
c_{m} \int_{\mathbb{R}^{m}} K(x, z) K(x, y) h(x)^{2} d \mu(x)=e^{(\nu(y)+\nu(z)) / 2} K(y, z)
$$

Proof. In Lemma 9.8, we have established

$$
e^{\nu(y) / 2} p(y)=c_{m} \int_{\mathbb{R}^{m}}\left(e^{-\Delta_{k} / 2} p(x)\right) K(x, y) h^{2}(x) d \mu(x),
$$

for polynomials. Fix $z \in \mathbb{C}^{m}$. Define $p_{j}(x)=\sum_{i=1}^{j} K_{j}(x, z)$. Then $p_{j}(y) \rightarrow K(y, z)$ and $e^{\Delta_{k} / 2} p_{j}(x) \rightarrow e^{-\nu(z) / 2} K(x, z)$ as $j \rightarrow \infty$. The result follows by dominated convergence.

Lemma 9.10. [12, Prop. 2.4] For For $j, l, n_{1}, n_{2} \in \mathbb{N}, p \in H_{k, 1}$ and $q \in H_{k, n_{2}}$, we have

$$
c_{m} \int_{\mathbb{R}^{m}} \phi_{j}(p ; x) \phi_{l}(q ; x) h^{2} d x=\delta_{j l} \delta_{n_{1} n_{2}} 2^{-\gamma-m / 2}(2 \pi)^{m / 2} \frac{(m / 2+\gamma)_{j+n}}{j} c_{m}^{\prime} \int p q h^{2} d \omega
$$

Proof. By using spherical coordinates the first integral is equal to

$$
c_{m} \frac{2^{m / 2-1}}{\Gamma(m / 2)} \int_{0}^{\infty} L_{j}^{n_{1}+\gamma+m / 2-1}\left(|x|^{2}\right) L_{l}^{n_{2}+\gamma+m / 2-1}\left(|x|^{2}\right) e^{-|x|^{2}}|x|^{n_{1}+n_{2}+2 \gamma+m-1} d|x| \int_{S} p q h^{2} d \omega
$$

Again the inner integral is zero unless $n_{1}=n_{2}$. Assuming $n_{1}=n_{2}$, we can substitute $t=|x|^{2}$ in the outer integral. By the orthogonality realition (9.1), we find for the outer integral

$$
\int_{0}^{\infty} L_{j}^{n_{1}+\gamma+m / 2-1}(t) L_{l}^{n_{1}+\gamma+m / 2-1}(t) e^{-t} t^{n_{1}+\gamma+m / 2-1} 1 / 2 d t=1 / 2 \delta_{j l} \frac{\Gamma\left(n_{1}+\gamma+m / 2+j\right)}{j!}
$$

and by combining the two integrals and using equation (6.6) we find the result of the Lemma.

Theorem 9.11. [12, Thm. 2.5] The linear span of $\left\{\phi_{j}(p): j, n \in \mathbb{N}, p \in H_{k, n}\right\}$ is dense in $L^{2}\left(\mathbb{R}^{m}, h^{2} d \mu\right)$.

Proof. See [12, p.129] for a proof by using Hamburger's Theorem.
Theorem 9.12. [12, Thm 2.6] For $j, n \in \mathbb{N}, p \in H_{k, n}, y \in \mathbb{R}^{m}, \phi_{m}(p)^{\wedge}(y)=(-1)^{n+2 j} \phi(p ; y)$ Proof. Denote $A=m / 2+n+\gamma-1$, then by Lemma 9.6 and Lemma 9.8 we can write

$$
(2 \pi)^{-m / 2} c_{m} \int_{\mathbb{R}^{m}} L_{j}^{A}\left(|x|^{2} / 2\right) p(x) K(x, y) h(x)^{2} e^{-|x|^{2} / 2} d x=(-1)^{j}\left(j!2^{j}\right)^{-1} e^{\nu(y) / 2} \nu(y)^{j} p(y)
$$

By using the identity

$$
L_{l}^{A}(t)=\sum_{j=0}^{l} 2^{j} \frac{(A+1)_{l}}{(A+1)_{j}} \frac{(-1)^{l-j}}{(l-j)!} L_{j}^{A}(t / 2),
$$

which is a special case of [23, problem 67, p.385], we can rewrite the equation above as

$$
e^{\nu(y) / 2} p(y)(-1)^{l} \frac{(A+1)_{l}}{l!} \sum_{j=0}^{l} \frac{(-l)_{j}}{(A+1)_{j}} \frac{(-\nu(y))^{j}}{j!} .
$$

Replace $y$ by $-i y\left(y \in \mathbb{R}^{n}\right)$, then $\nu(y)$ becomes $-|y|^{2}$ and $p(y)$ becomes $(-i)^{n} p(y)$ and the sums yields a Laguerre polynomial.
The integral becomes equal to

$$
(-1)^{m}(-i)^{n} p(y) L_{m}^{A}\left(|y|^{2}\right) e^{-|y|^{2} / 2}
$$

Corollary 9.13. [12, Cor 2.7] The Dunkl transform has period 4 and extends to an isometry from $L^{2}\left(\mathbb{R}^{m}, h^{2} d x\right)$ onto itself. The square of the transform is the central involution, that is, if $\left(\mathscr{D}_{k} f\right)(x)=g(x)$, then $\left(\mathscr{D}_{k} g\right)(x)=f(-x)$ almost everywhere.

Lemma 9.14. [12, Lemma 2.9] Let $f \in \mathbb{E}\left(\mathbb{R}^{m}\right)$ and let $g \in C^{\infty}\left(\mathbb{R}^{m}\right)$ such that $g$ and all its partial derivatives are $O(\exp (B|x|)$ for some $B<\infty$. This includes $g(x)=K(x, y)$ for fixed $y$. Then

$$
\int_{\mathbb{R}^{m}}\left(T_{j} f\right) g h^{2} d x=-\int_{\mathbb{R}^{m}}\left(f T_{j}\right) g h^{2} d x
$$

Proof. To prove this, we need to use integration by parts. This is possible since $f$ and $g$ decrease rapidly at infinity. At first we require $k(\alpha)>1(\alpha \in R)$, so $1 /\langle x, \alpha\rangle$ is integrable for $h^{2} d x$. After the result is established, we can drop this restriction (back at $k(\alpha)>0(\alpha \in R)$ ) by analytic continuation. Now

$$
\begin{aligned}
\int_{\mathbb{R}^{m}} & \left(T_{j} f\right) g h^{2} d x= \\
& -\int_{\mathbb{R}^{m}} f(x) \frac{\partial}{\partial x_{j}}\left(g(x) h^{2}(x)\right) d x \\
& +\sum_{\alpha \in R_{+}} k_{\alpha} \alpha_{j} \int_{\mathbb{R}^{m}} \frac{f(x)-f\left(r_{\alpha}(x)\right)}{\langle x, \alpha\rangle} g(x) h^{2}(x) d x \\
= & -\int_{\mathbb{R}^{m}}\left[f(x) \frac{\partial}{\partial x_{j}}\left(g(x)+2 f(x) \sum_{\alpha \in R_{+}} k_{\alpha} \frac{\alpha_{j}}{\langle x, \alpha\rangle} g(x)\right] h^{2}(x) d x\right. \\
& +\sum_{\alpha \in R_{+}} k_{\alpha} \alpha_{j} \int_{\mathbb{R}^{m}} f(x) \frac{g(x)+g\left(r_{\alpha}(x)\right)}{\langle x, \alpha\rangle} h^{2}(x) d x \\
= & -\int_{\mathbb{R}^{m}} f\left(T_{j}(g)\right) h^{2} d x,
\end{aligned}
$$

where the substitution $x \rightarrow r_{\alpha} x$, for which $\langle x, \alpha\rangle$ becomes $\left\langle r_{\alpha} x, \alpha\right\rangle=\left\langle x, r_{\alpha} \alpha\right\rangle=-\langle x, \alpha\rangle$, was used to show that

$$
\int_{\mathbb{R}^{m}} \frac{f\left(r_{\alpha}(x)\right) g(x)}{\langle x, \alpha\rangle} h^{2}(x) d x=-\int_{\mathbb{R}^{m}} \frac{f(x) g\left(r_{\alpha}(x)\right)}{\langle x, \alpha\rangle} h^{2}(x) d x .
$$

Theorem 9.15. [12, Thm 2.10] For $f \in E\left(\mathbb{R}^{m}\right)$, we have that $\left(T_{j} f\right)^{\wedge}(y)=i y_{j}\left(\mathscr{D}_{k} f\right)(y)$ The operator $-i T_{j}$ is densely defined on $L^{2}\left(\mathbb{R}^{m}, h^{2} d x\right)$ and is self-adjoint.

Proof. For fixed $y \in \mathbb{R}^{m}$, put $g(x)=K(x,-i y)$ in Lemma 9.14. Then $T_{j} g(x)=-i y_{j} K(x,-i y)$ and $\mathscr{D}_{k}\left(T_{j} f\right)(y)=(-1)\left(-i y_{j}\right)\left(\mathscr{D}_{k} f\right)(y)$. The multiplication operator defined by $M_{j} f(y)=$ $y_{j} f(y)$ is densely defined and self-adjoint on $L^{2}\left(\mathbb{R}^{m}, h^{2} d x\right)$. Further $-i T_{j}$ is the inverse image of $M_{j}$ under the Dunkl transform, an isometric isomorphism.

Corollary 9.16. [12, Cor. 2.11] For $f \in E\left(\mathbb{R}^{m}\right)$, define $g_{j}=x_{j} f(1 \leq j \leq m)$. The Dunkl transform of $g_{j}$ is given by

$$
\mathscr{D}_{k}\left(g_{j}\right)(y)=i T_{j}\left(\mathscr{D}_{k} f\right)(y), \quad\left(y \in \mathbb{R}^{m}\right) .
$$

## Chapter 10

## Use of the Dunkl transformation in differential-difference equations

The Dunkl operators are generalizations of the partial derivatives. It makes sense to use them to generalize differential equations. In this chapter we shall look at the generalized heat equation. To do this we shall generalize Section 4.3 .1 of [15] to the setting of Dunkl operators.

Definition 10.1. The Fourier transform of a function $f \in \mathcal{S}\left(\mathbb{R}^{m}\right)$ is given by

$$
\hat{f}(y)=\frac{1}{(2 \pi)^{m / 2}} \int_{\mathbb{R}^{m}} f(x) e^{-i\langle x, y\rangle} d x\left(\in \mathcal{S}\left(\mathbb{R}^{m}\right)\right)
$$

The inverse Fourier transform of a function $f \in \mathcal{S}\left(\mathbb{R}^{m}\right)$ is given by

$$
\check{f}(y)=\frac{1}{(2 \pi)^{m / 2}} \int_{\mathbb{R}^{m}} f(x) e^{i\langle x, y\rangle} d x\left(\in \mathcal{S}\left(\mathbb{R}^{m}\right)\right)
$$

Here $\mathcal{S}\left(\mathbb{R}^{m}\right)$ is the Schwartz-space, defined by

$$
\left\{f \in C^{\infty}\left(\mathbb{R}^{m}:\left|\partial^{\alpha} x^{\beta} f\right|<\infty, \text { for all multi-indices } \alpha, \beta\right\}\right.
$$

Since the Fourier transform is an isometric isomorphism on $\mathcal{S}\left(\mathbb{R}^{m}\right)$ and $\mathcal{S}\left(\mathbb{R}^{m}\right)$ is dense in $L^{2}\left(\mathbb{R}^{m}\right)$, the Fourier transform and its inverse can be extended to all of $L^{2}\left(\mathbb{R}^{m}\right)$ in the following way; for $f \in L^{2}\left(\mathbb{R}^{m}\right)$ take a sequence $\left(f_{n}\right)_{n \in \mathbb{N}} \in \mathcal{S}\left(\mathbb{R}^{m}\right)$ converging to $f$. Then the sequence $\hat{f}_{n}$ converges to an element $g \in L^{2}\left(\mathbb{R}^{m}\right)$. Note that $g$ only depends on the choice of $f$ and not the choice of the converging sequence. We define $g=\hat{f}$.
In the following we will also use the Dunkl transform, which was defined in Definition 9.2 as

$$
\left(\mathscr{D}_{k} f\right)(y)=(2 \pi)^{m / 2} c_{m} \int_{\mathbb{R}^{m}} f(x) K(x,-i y)|h(x)|^{2} d x
$$

with inverse given by

$$
\left(\mathscr{D}_{k}^{-1} f\right)(y)=(2 \pi)^{m / 2} c_{m} \int_{\mathbb{R}^{m}} f(x) K(x, i y)|h(x)|^{2} d x
$$

for $\left(f \in \mathbb{E}\left(\mathbb{R}^{m}\right)\right.$. We have the relation $\mathscr{D}_{k} f(-y)=\mathscr{D}_{k}^{-1} f(y)$. By Corollary 9.13 the Dunkl transform extends to a transform on $L^{2}\left(\mathbb{R}^{m}, h^{2} d x\right)$. This extension is defined in the following
way; for $f \in L^{2}\left(\mathbb{R}^{m}, h^{2} d x\right)$ take a sequence $\left.\left(f_{n}\right)_{n \in \mathbb{N}}\right) \in \mathbb{E}\left(\mathbb{R}^{m}\right)$ converging to $f$. Then the sequence $\hat{f}_{n}$ converges to an element $g \in L^{2}\left(\mathbb{R}^{m}, h^{2} d x\right)$. Note that $g$ only depends on the choice of $f$ and not the choice of the converging sequence. We define $g=\mathscr{D}_{k}(f)$.

Recall that $c_{k}^{-1}$ is given by $\int_{\mathbb{R}^{m}}(2 \pi)^{-m / 2}|h(x)|^{2} e^{-|x|^{2}}$. To simplify the notation it is useful to define

$$
\zeta_{m}=(2 \pi)^{-m / 2} c_{m}=\left(\int_{\mathbb{R}^{m}}|h(x)|^{2} e^{-|x|^{2}}\right)^{-1}
$$

The Fourier transform can be used to simplify certain types of differential equations. Consider as example the Cauchy problem defined by

$$
\begin{cases}p\left(\frac{\partial}{\partial x}\right) u_{0}-\frac{\partial}{\partial t} u_{0}=0 & \text { on } \mathbb{R}^{m} \times(0, \infty),  \tag{10.1}\\ u_{0}=f & \text { on } \mathbb{R}^{m} \times(t=0) .\end{cases}
$$

Here we assume that $f(x) \in C\left(\mathbb{R}^{m}\right)_{0}$, the space of continuous functions with compact support, and we search for a solution $u(\cdot, t) \in L^{2}\left(\mathbb{R}^{m}\right)$, for all $t>0$. Also $p\left(\partial / \partial_{x}\right)$ is obtained by replacing $x_{i}$ with $\partial_{i}$ in the expression of $p(x)$, as was defined in Definition 5.1.

Lemma 10.2. The solution for the Cauchy problem in equation (10.1) is given by the double integral

$$
u_{0}(x, t)=\frac{1}{(2 \pi)^{m}} \int_{\mathbb{R}^{m}}\left[\int_{\mathbb{R}^{m}} \exp (p(i \xi) t) \exp \langle i(x-y), \xi\rangle d \xi\right] f(y) d y,
$$

which should be interpreted in the sense of distribution theory.
Proof. To solve this problem, we apply the steps in [15, p, 188]. First, we apply the Fourier transform to the Cauchy problem and obtain the resulting system

$$
\frac{\partial \hat{u}_{0}}{\partial t}=p(i x) \hat{u}_{0}, \hat{u}_{0}(x, 0)=\hat{f}(x),
$$

which has the solution $\hat{u}_{0}=\hat{f} \exp (p(i x) t)$. Applying the inverse Fourier transform to this result gives the solution

$$
u_{0}(x, t)=\frac{\exp (p(i x) t)^{2} * f(x)}{(2 \pi)^{m / 2}},
$$

which is equal to the double integral

$$
u_{0}(x, t)=\frac{1}{(2 \pi)^{m}} \int_{\mathbb{R}^{m}}\left[\int_{\mathbb{R}^{m}} \exp (p(i \xi) t) \exp \langle i(x-y), \xi\rangle d \xi\right] f(y) d y .
$$

By taking the derivatives into the integral we see that $u_{0}(x, t)$ is indeed a solution and by setting $t=0$, we are left with

$$
u_{0}(x, t)=\frac{1}{(2 \pi)^{m / 2}} \int_{\mathbb{R}^{m}}\left[\int_{\mathbb{R}^{m}} \exp \langle i(x-y), \xi\rangle d \xi\right] f(y) d y=\int_{\mathbb{R}^{m}} \delta(x-y) f(y) d(y)=f(x),
$$

so $u_{0}(x, t)$ satisfies the initial condition.
Next we can replace the partial derivatives with Dunkl operators to get the system

$$
\begin{cases}P\left(T^{x}\right) u_{k}-\frac{\partial}{\partial t} u_{k}=0 & \text { on } \mathbb{R}^{m} \times(0, \infty),  \tag{10.2}\\ u_{k}=f & \text { on } \mathbb{R}^{m} \times(t=0),\end{cases}
$$

again for $u_{k}(\cdot, t) \in L^{2}\left(\mathbb{R}^{m}\right), \forall t>0$ and $f \in C(\mathbb{R})_{0}$.

Lemma 10.3. The solution for the Cauchy problem in equation (10.2) is given by the double integral

$$
u_{k}(x, t)=\zeta_{m} \int_{\mathbb{R}^{m}} \int_{\mathbb{R}^{m}} \exp (P(i \xi) t) f(y) K(i x, \xi) K(-i y, \xi)|h(\xi)|^{2}|h(y)|^{2} d \xi d y
$$

Proof. This Cauchy problem was solved in [21, Ch. 4] for $P=|x|^{2}$. In this proof we shall generalize the method of [21] to an arbitrary polynomial.
Since $T_{i}^{x} K(x, y)=y_{i} K(x, y)$ we can simplify this system by applying the Dunkl transform. This gives the system

$$
P(i x) \mathscr{D}_{k}\left(u_{k}\right)=\frac{\partial}{\partial t} \mathscr{D}_{k}\left(u_{k}\right), \quad \mathscr{D}_{k} u_{k}(x, 0)=\mathscr{D}_{k}(f)(x),
$$

with the solution

$$
\mathscr{D}_{k}\left(u_{k}\right)=\mathscr{D}_{k}(f) \exp (P(i x) t) .
$$

The solution of the original system can be found by applying the inverse Dunkl transform to both sides. But in this general situation it is not known whether there exists a reasonable convolution structure on $\mathbb{R}^{m}$ matching the action of the Dunkl transform $\mathscr{D}_{k}$.
However, we can find a solution by using the generalized translation [21, Eqn. (4.2)], which is defined by

$$
L_{k}^{y} f(x)=\zeta_{m} \int_{\mathbb{R}^{m}}\left(\mathscr{D}_{k} f\right)(\xi) K(i x, \xi) K(i y, \xi)|h(\xi)|^{2} d \xi
$$

Note that $L_{k}^{y} f(x)=L_{k}^{x} f(y), L_{k}^{0} f(x)=f(x)$ and $L_{0}^{y} f(x)=f(x+y)$.
Define $F \in L^{2}\left(\mathbb{R}^{m}, h^{2} d x\right)$ by

$$
\left(\mathscr{D}_{k} F\right)(x)=(\exp (P(i x) t)) .
$$

By using the generalized translation as convolution structure, we find the solution

$$
u_{k}(x, t)=\int_{\mathbb{R}^{m}} L_{k}^{-y} F(x) f(y)|h(y)|^{2} d y
$$

which is equal to the double integral

$$
\zeta_{m} \int_{\mathbb{R}^{m}} \int_{\mathbb{R}^{m}} \exp (P(i \xi) t) f(y) K(i x, \xi) K(-i y, \xi)|h(\xi)|^{2}|h(y)|^{2} d \xi d y
$$

By taking the Dunkl operators into the integral, which is justified because the integrands decreases rapidly to 0 at infinity, we see that

$$
\begin{aligned}
& \left(P(T)-\frac{\partial}{\partial t}\right) \zeta_{m} \int_{\mathbb{R}^{m}} \int_{\mathbb{R}^{m}} \exp (P(i \xi) t) f(y) K(i x, \xi) K(-i y, \xi)|h(\xi)|^{2}|h(y)|^{2} d y \\
& \quad=\zeta_{m} \int_{\mathbb{R}^{m}} \int_{\mathbb{R}^{m}}(P(i \xi)-P(i \xi)) \exp (P(-i x) t) f(y) K(i x, \xi) K(-i y, \xi)|h(\xi)|^{2}|h(y)|^{2} d y \\
& \quad=0,
\end{aligned}
$$

which shows that $u_{k}(t, x)$ is the solution we were looking for.

We can also look at the Cauchy problem, which is obtained by applying the intertwining operator $V_{k}^{x}$ to equation (10.1). Since the intertwining operator leaves the $t$-variable invariant we get

$$
\begin{cases}P\left(T^{x}\right) V_{k} u_{0}-\frac{\partial}{\partial t} V_{k} u_{0}=0 & \text { on } \mathbb{R}^{m} \times(0, \infty),  \tag{10.3}\\ V_{k} u_{0}=V_{k} f & \text { on } \mathbb{R}^{m} \times(t=0),\end{cases}
$$

This is precisely the Cauchy problem in Equation 10.2 with initial condition $u_{k}(x, 0)=V_{k}(f)$. We can also solve this problem by applying $V_{k}^{x}$ to the solution of the Cauchy problem in equation (10.1) with initial value $u_{0}(x, 0)=f(x)$. To compare the two solutions we need to consider equation 10.2 with initial condition $V_{k}\left(u_{k}(x, 0)\right)=f(x)=V_{k} V_{k}^{-1} f(x)$. This gives

$$
\begin{aligned}
u_{k}^{\prime}(x, t) & =\frac{1}{(2 \pi)^{m}} V_{k}^{x}\left(\int_{\mathbb{R}^{m}}\left[\int_{\mathbb{R}^{m}} \exp (P(i \xi) t) \exp \langle i(x-y), \xi\rangle d \xi\right]\left(V_{k}^{y}\right)^{-1} f(y) d y\right), \\
& \left.=\frac{1}{(2 \pi)^{m}} \int_{\mathbb{R}^{m}}\left[\int_{\mathbb{R}^{m}} \exp (P(i \xi) t) K(i x, \xi) \exp \langle-i y), \xi\right\rangle d \xi\right]\left(V_{k}^{y}\right)^{-1} f(y) d y
\end{aligned}
$$

This $u_{k}^{\prime}(x, t)$ must be equal to the $u_{k}(x, t)$ in Lemma 10.3 since both functions solve the Cauchy problem (10.2) for the same initial value. This gives the equation

$$
\begin{aligned}
& \zeta_{m} \int_{\mathbb{R}^{m}} \int_{\mathbb{R}^{m}} \exp (P(i \xi) t) K(i x, \xi) K(-i y, \xi)|h(\xi)|^{2} d \xi f(y)|h(y)|^{2} d y \\
& \quad=\frac{1}{(2 \pi)^{m}} \int_{\mathbb{R}^{m}}\left[\int_{\mathbb{R}^{m}} \exp (P(i \xi) t) K(i x, \xi) \exp \langle-i y, \xi\rangle d \xi\right]\left(V_{k}^{y}\right)^{-1} f(y) d y
\end{aligned}
$$

This gives some information about the intertwining operator. This is not enough to define the intertwining operator in a closed form, since the Dunkl transform contains the intertwining operator acting on $\exp \langle-i x, y\rangle$.

As an example we can apply this method to the Dunkl heat equation, where $P(T)=\Delta_{k}$ and the Cauchy problem is given by

$$
\begin{cases}\Delta_{k} u_{k}-\frac{\partial}{\partial t} u=0 & \text { on } \mathbb{R}^{m} \times(0, \infty),  \tag{10.4}\\ u_{k}=f & \text { on } \mathbb{R}^{m} \times(t=0),\end{cases}
$$

Lemma 10.4. [21, Thm. 4.11] The solution of the Cauchy problem (10.4) is given by

$$
u_{k}(x, t)=\int_{\mathbb{R}^{m}} \frac{1}{(4 t)^{\gamma+m / 2}} e^{-\left(|x|^{2}+|y|^{2}\right) / 4 t} K\left(\frac{x}{\sqrt{2 t}}, \frac{y}{\sqrt{2 t}}\right)|h(y)|^{2} f(y) d y .
$$

Proof. First of all we need to apply the inverse Dunkl transform to the function $e^{-|\xi|^{2} t}$. For this, we use Definition 9.4 , to write $\phi_{0}(1 ; x)=1 \cdot L_{0}^{\gamma+m / 2-1}\left(|x|^{2}\right) e^{|x|^{2} / 2}=e^{-|x|^{2} / 2}$ and apply Theorem 9.12 , to get $\mathscr{D}_{k} e^{-|x|^{2} / 2}=e^{-|x|^{2} / 2}$.
By setting $x=\sqrt{2 t} \xi$, we can see that

$$
\begin{aligned}
\int e^{-t|x|^{2}}|h(x)|^{2} K(-i x, y) d x & =\int e^{-|\xi|^{2} / 2} h^{2}(\xi / \sqrt{2 t}) K(-i \xi / \sqrt{2 t}, y) d(\xi / \sqrt{2 t}), \\
& =\int e^{-|\xi|^{2} / 2}|h(x)|^{2} d(\xi) K(-i \xi, y / \sqrt{2 t})(\sqrt{2 t})^{-m-2 \gamma}, \\
& =e^{-|y|^{2} /(4 t)}(\sqrt{4 t})^{-m-2 \gamma}
\end{aligned}
$$

So

$$
e^{-|\xi|^{2} t}=\mathscr{D}_{k}\left(\frac{1}{(4 t)^{\gamma+m / 2}} \exp \left(-\frac{|\xi|^{2}}{4 t}\right)\right)
$$

By applying the previous method we find as solution

$$
u_{k}(x, t)=\zeta_{m} \int_{\mathbb{R}^{m}} \int_{\mathbb{R}^{m}} \exp \left(-|\xi|^{2} t\right) f(y) K(i x, \xi) K(-i y, \xi)|h(\xi)|^{2}|h(y)|^{2} d \xi d y
$$

By Lemma (9.9) on page 71 we have

$$
\zeta_{m} \int_{\mathbb{R}^{m}} K(y, x) K(z, x)|h(x)|^{2} e^{-|x|^{2} / 2} d x=e^{\nu(x)+\nu(y) / 2} K(y, z),
$$

which can be used to solve the $d \xi$ integral.
By using the substitution $\sqrt{2 t} \xi=\eta$, we get

$$
\begin{aligned}
& \zeta_{m} \int_{\mathbb{R}^{m}} e^{-|\xi|^{2} t} K(i x, \xi) K(-i y, \xi)|h(\xi)|^{2} d \xi \\
&=\zeta_{m} \int_{\mathbb{R}^{m}} e^{-|\eta|^{2} / 2} K\left(i x, \frac{\eta}{\sqrt{2 t}}\right) K\left(-i y, \frac{\eta}{\sqrt{2 t}}\right)\left|h\left(\frac{\eta}{\sqrt{2 t}}\right)\right|^{2} d^{m}\left(\frac{\eta}{\sqrt{2 t}}\right), \\
&=\frac{\zeta_{m}}{(4 t)^{\gamma+m / 2}} \int_{\mathbb{R}^{m}} e^{-|\eta|^{2} / 2} K\left(\frac{i x}{\sqrt{2 t}}, \eta\right) K\left(\frac{-i y}{\sqrt{2 t}}, \eta\right)|h(\eta)|^{2} d \eta \\
&=\frac{1}{4 t)^{\gamma+m / 2}} e^{-\left(|x|^{2}+|y|^{2}\right) / 4 t} K\left(\frac{i x}{\sqrt{2 t}}, \frac{-i y}{\sqrt{2 t}}\right) \\
& \quad=\frac{1}{(4 t)^{\gamma+m / 2}} e^{-\left(|x|^{2}+|y|^{2}\right) / 4 t} K\left(\frac{x}{\sqrt{2 t}}, \frac{y}{\sqrt{2 t}}\right)
\end{aligned}
$$

which gives the solution

$$
u_{k}(x, t)=\int_{\mathbb{R}^{m}} \frac{1}{(4 t)^{\gamma+m / 2}} e^{-\left(|x|^{2}+|y|^{2}\right) / 4 t} K\left(\frac{x}{\sqrt{2 t}}, \frac{y}{\sqrt{2 t}}\right)|h(y)|^{2} f(y) d y .
$$

Denote the space $\mathbb{R}^{m} \times(0, t)$ by U. Then, according to [21, p. $536-540$ ], this is the unique solution within the class of function $C^{2}(U) \cap C(\bar{U})$, which satisfy the following exponential growth condition: There exist positive constants $C, \lambda$ and $r$, such that

$$
\left|u_{k}(x, t)\right| \leq C \cdot e^{-\lambda|x|^{2}}, \text { for all }(x, t) \in U \text { with }|x|>r .
$$

Definition 10.5. [21, Def. 4.6] The generalized heat kernel $\Gamma_{k}(x, y, t): \mathbb{R}^{m} \times \mathbb{R}^{m} \times \mathbb{R}_{+} \rightarrow \mathbb{R}$ is given by

$$
\Gamma_{k}(x, y, t)=\frac{1}{(4 t)^{\gamma+m / 2}} \exp \left(-\frac{|x|^{2}+|y|^{2}}{4 t}\right) K\left(\frac{x}{\sqrt{2 t}}, \frac{y}{\sqrt{2 t}}\right) .
$$

We can use this to write the solution of the Dunkl heat equation as

$$
u(x, t)=\int_{\mathbb{R}^{m}} \Gamma_{k}(x, y, t) f(y)\left|h^{2}(y)\right| d y .
$$

As before, we can find another way to write this solution by applying $V_{k}$ to the solution of the Cauchy problem of the ordinary heat equation.

The solution for the Cauchy problem of the heat equation is given by

$$
u_{0}(x, t)=\frac{1}{(4 \pi t)^{m / 2}} \int_{\mathbb{R}^{m}} \exp \left(\frac{-|x-y|^{2}}{4 t}\right) f(y) d y,
$$

and the solution of equation with initial condition $u_{0}(x, 0)=V_{k}^{-1} f(x)$ is given by

$$
\begin{aligned}
u_{k}(x, t) & =V_{k}^{x}\left(\frac{1}{(4 \pi t)^{m / 2}} \int_{\mathbb{R}^{m}} \exp \left(\frac{-|x-y|^{2}}{4 t}\right)\left(V_{k}^{-1} f\right)(y) d y\right) \\
& =\frac{1}{(4 \pi t)^{m / 2}} \int_{\mathbb{R}^{m}} V_{k}^{x}\left(\exp \left(\frac{-|x-y|^{2}}{4 t}\right)\right)\left(V_{k}^{-1} f\right)(y) d y \\
& =\frac{1}{(4 \pi t)^{m / 2}} \int_{\mathbb{R}^{m}} V_{k}^{x}\left(\exp \left(-\frac{|x|^{2}+|y|^{2}}{4 t}\right) \exp \left\langle\frac{x}{\sqrt{2 t}}, \frac{y}{\sqrt{2 t}}\right\rangle\right)\left(V_{k}^{-1} f\right)(y) d y
\end{aligned}
$$

which again might contain useful information to determine the closed form of the intertwining operator.
For the root system $A_{m-1}$, we can simplify the solution of equation 10.4 given in Lemma 10.4

Denote by $\theta_{m}(x)$ the $m$-dimensional Vandermonde determinant. Then by [1, p.24] the value of $c_{m}$ is given by

$$
\begin{aligned}
c_{m}^{-1} & =\int_{\mathbb{R}^{m}}|h(x)|^{2 k} e^{-|x|^{2}} d x \\
& =2^{-k m(m+1) / 2} 2^{-\gamma-m / 2} \int_{\mathbb{R}^{m}} e^{-|x|^{2} / 2}\left|\theta_{m}(x)\right|^{2 k} d x \\
& =2^{-2 \gamma}(\pi)^{m / 2} \prod_{j=1}^{m} \frac{\Gamma(1+j k)}{\Gamma(1+j)},
\end{aligned}
$$

which leads to

$$
\begin{align*}
& V_{k}^{x}\left(e^{-\left(|x|^{2}+|y|^{2}\right) / 4 t} e^{\langle x, y\rangle / 2 t}\right) \\
& \quad=4^{\gamma} \prod_{j=1}^{m} \frac{\Gamma(1+j k)}{\Gamma(1+j)}\left|\theta_{m}\left(\frac{y}{\sqrt{2 t}}\right)\right|^{2} e^{-\left(|x|^{2}+|y|^{2}\right) / 4 t} V_{k}^{x} e^{\langle x, y\rangle / 2 t}, \tag{10.5}
\end{align*}
$$

and setting $k=0$, which means $\gamma=0$ and $h=1$ gives

$$
V_{0}^{x}\left(e^{-\left(|x|^{2}+|y|^{2}\right) / 4 t} e^{\langle x, y\rangle / 2 t}\right)=e^{-\left(|x|^{2}+|y|^{2}\right) / 4 t} V_{0}^{x} e^{\langle x, y\rangle / 2 t},
$$

as expected, since $V_{0}$ is the identity.

## Chapter 11

## Application of Dunkl operators in physics

The Dunkl operators occur in a natural way in the study of certain types Calogero-MoserSutherland models or CMS models. Basically a CMS model is a quantum mechanical model of $m$ particles moving on a line or circle, under influence of some two body interactions and an external potential.
For some theory about this type of models we will follow [19], to construct a set of coupled momentum operators $\pi_{i}$. These operators are a gauge-transformed form of the Dunkl operators $T_{i}$ of type $A_{n}$. By a result in [19], this physical system is integrable and its solution is related to the Dunkl heat equation.
In the following all indices $i, j, l$ will run from 1 to $m$. Also when $i$ is not an index, it will be used as the complex unit element with $i^{2}=-1$.
We will only look at a quantum mechanical model of $m$ particles moving on the real line, with positions given by $x_{i}(1 \leq i \leq m)$ and momenta given by $p_{i}(1 \leq i \leq m)$. The coordinates are canonical so $\left[x_{i}, p_{j}\right]=\delta_{i j}$.
For an arbitrary potential $V(x): \mathbb{R} \rightarrow \mathbb{C}$, we define the coupled momentum operators by

$$
\pi_{i}=p_{i}+i \sum_{i \neq j} M_{i j} V_{i j}
$$

where $V_{i j}=V\left(x_{i}-x_{j}\right)$ and $M_{i j}$ is the particles permutation operator, which obeys

$$
M_{i j}^{2}=1, M_{i j}=M_{j i}=M_{i j}^{\dagger}
$$

and

$$
M_{i} j B_{j}=B_{i} M_{i} j, \quad M_{i j} B_{k}=B_{k} M_{i j}, k \neq i, k \neq j,
$$

where $B_{i}$ can be any operator carrying an particle index. Here $M_{i j}^{\dagger}$ denotes the hermitian adjoint of $M_{i j}$. By looking at the root system $R=A_{m-1}$ we can rewrite the momenta as

$$
\pi_{i}=p_{i}+i \sum_{\alpha \in \mathbb{R}^{+}} r_{\alpha} V_{\alpha},
$$

with $V_{\alpha}=V(\langle x, \alpha\rangle)$. We consider the Hamiltonian which takes a free form in terms of $\pi_{i}$, given by

$$
H=\frac{1}{2} \sum_{i=1}^{m} \pi_{i}^{2} .
$$

We want to impose the Hermiteness condition $\pi_{i}=\pi_{i}^{\dagger}$ to make sure the momenta are real. Then

$$
\begin{aligned}
p_{i}+i \sum_{i \neq j} V_{i j} M_{i j} & =p_{i}^{\dagger}-i \sum_{i \neq j}^{m} V_{i j}^{\dagger} M_{i j}^{\dagger}, \\
\sum_{i \neq j} V_{i j} M_{i j} & =-\sum_{i \neq j}^{m} V_{i j}^{\dagger} M_{i j}, \\
V\left(x_{i}-x_{j}\right) & =-\overline{V\left(x_{j}-x_{i}\right)}, \\
V(x)^{\dagger} & =-V(-x),
\end{aligned}
$$

where $\bar{f}$ is the complex conjugate of $f$ and is used that $\overline{V(x)}=V(x)^{\dagger}$, because $V: \mathbb{R} \rightarrow \mathbb{C}$. To simplify the notation we write $V_{i j l}=V_{i j} V_{i l}+V_{j l} V_{i l}+V_{i j} V_{j l}$ and denote the generator of cyclic permutations in three indices by $M_{i j l}=M_{i l} M_{j l}$. Note that

$$
\begin{aligned}
\sum_{i \neq j} \sum_{l \neq s} M_{i j} M_{l s} V_{i j} V_{j k} & =\sum_{\substack{,, j, l, s \\
\text { different }}} M_{i j} M_{l s} V_{i j}\left(V_{l s}+V_{s l}\right)+\sum_{i \neq j \neq l \neq i} M_{i j} M_{j l} V_{i j} V_{j l}+\sum_{i \neq j} M_{i j}^{2} V_{i j}^{2} \\
& =\frac{1}{2} \sum_{\substack{i, j, l, s \\
\text { different }}} M_{i j} M_{l s} V_{i j} V_{l s}+\frac{1}{3} \sum_{i \neq j \neq l \neq i} M_{i j l} V_{i j l}+\sum_{i \neq j} V_{i j}^{2} \\
& =\frac{1}{3} \sum_{i \neq j \neq l \neq i} M_{i j l} V_{i j l}+\sum_{i \neq j} V_{i j}^{2}
\end{aligned}
$$

We can use this to rewrite this Hamiltonian in the coordinates $x_{i}$ and $p_{i}$, which gives

$$
H=\frac{1}{2} \sum_{i=1}^{m} \pi_{i}^{2}=\frac{1}{2} \sum_{i=1}^{m} p_{i}^{2}+\frac{1}{2} \sum_{i \neq j}\left[i V_{i j}\left(p_{i}+p_{j}\right) M_{i j}+V_{i j}^{\prime} M_{i j}+V_{i j}^{2}\right]-\frac{1}{6} \sum_{i \neq j \neq l \neq i} V_{i j l} M_{i j l},
$$

by a straightforward calculation.
By looking at the root system $A_{n-1}$, we can rewrite this as

$$
\begin{equation*}
H=\frac{1}{2} \sum_{i=1}^{m} \pi_{i}^{2}=\frac{1}{2} \sum_{i=1}^{m} p_{i}^{2}+\frac{1}{2} \sum_{\alpha \in R_{+}}\left[i V_{\alpha}\left(p_{i}+p_{j}\right) r_{\alpha}+V_{\alpha}^{\prime} r_{\alpha}+V_{\alpha}^{2}\right]-\frac{1}{2} \sum_{\substack{\alpha, \beta \in R_{+} \\ \alpha \neq \beta}} V_{\alpha} V_{\beta} r_{\alpha} r_{\beta} . \tag{11.1}
\end{equation*}
$$

We would like the Hamiltonian to contain a sum of kinetic and potential terms. This is achieved if $V(-x)=-V(x)$, since the terms linear in $p_{i}$ drop out. Also, we would like the Hamiltonian only to contain 2-body potentials. This gives the restriction

$$
V(x) V(y)+V(y) V(z)+V(z) V(x)=W(x)+W(y)+W(z),
$$

for $\mathrm{x}+\mathrm{y}+\mathrm{z}=0$, where $\mathrm{W}(\mathrm{x})$ is a new symmetric function. After these restriction we can write

$$
H=\frac{1}{2} \sum_{i=1}^{m} p_{i}^{2}+\sum_{i<j}\left[V_{i j}^{2}+V_{i j}^{\prime} M_{i j}+W_{i j} \sum_{i \neq k} M_{i j k}\right],
$$

and we have the commutator relation

$$
\begin{equation*}
\left[\pi_{i}, \pi_{j}\right]=\sum_{k \neq i, j} V_{i j k}\left[M_{i j k}-M_{j i k}\right] \tag{11.2}
\end{equation*}
$$

Consider $V(x)=k / x, k \in \mathbb{R}$. Since $\frac{1}{x_{i} x_{j}}+\frac{1}{x_{i} x_{l}}+\frac{1}{x_{j} x_{l}}=0$, for $x_{i}+x_{j}+x_{l}=0$, we find $W(x)=0$. This gives the Hamiltonian

$$
H=\frac{1}{2} \sum_{i=1}^{m} \pi_{i}^{2}=\frac{1}{2} \sum_{i=1}^{m} p_{i}^{2}+\frac{1}{2} \sum_{i \neq j}-\left[\frac{k s_{i j}}{\left(x_{i}-x_{j}\right)^{2}}+\frac{k^{2}}{\left(x_{i}-x_{j}\right)^{2}}\right]=H_{c m s}
$$

with the associated momentum operators

$$
\pi_{i}=i \tilde{T}_{i}=i \frac{\partial}{\partial x_{i}}+i \sum_{i=1, i \neq j}^{m} k \frac{M_{i j}}{x_{j}-x_{i}}
$$

From equation 11.2 it can be seen that these momenta commute. This means that the operators $I_{n}=\sum_{i} \pi_{i}^{n}, 1 \leq n \leq m$ are $m$ commuting conserved quantities, which shows that this system is integrable (see [19]).
The operators $\pi_{i}$ are not well-behaving around 0 for functions in $L^{2}\left(\mathbb{R}^{m}, d x\right)$, but the operators are well-behaving in the normed space $L^{2}\left(\mathbb{R}^{m},|h(x)|^{2} d x\right)$ (see [22, Ch.3.1]). We can modify the system by using the transform $f \rightarrow|h(x)| f$, which leads to the transformed hamiltonian

$$
\left.\bar{H}=|h(x)| H_{c m s}|h(x)|^{-1}=|h(x)|\left(-\frac{1}{2} \Delta+\frac{1}{2} \sum_{i \neq j}-\frac{k s_{i j}}{\left(x_{i}-x_{j}\right)^{2}}+\frac{k^{2}}{\left(x_{i}-x_{j}\right)^{2}}\right]\right)|h(x)|^{-1},
$$

and by a direct computation it can be seen that $\bar{H}=-\Delta_{k}$, which shows that the Dunkl operators occur in a natural way in some physical systems. The Schrödinger equation for this Hamiltonian can be solved by replacing $t$ with it in the solution for the Dunkl heat equation. Finally note that the right hand side of equation (11.1) is valid for any root system, although it is harder to find a physical meaning for these Hamiltonians.
See for example [6] and [22] for more results about the CMS-models in context of Dunkl operators.

## Chapter 12

## Dunkl processes

In this chapter we are going to have a look at certain stochastic processes involving Dunkl operators. We can use these processes to describe the action of the intertwining operator on symmetric polynomials. This involves some theory about Jack polynomials and hypergeometric functions, which will be explained along the way.

The ordinary heat equation is the Kolgomorov Backward Equation (KBE) of the Brownian motion, which is an example of a Markov process. We want to define Dunkl process, as the Markov process which has the Dunkl Heat Equation as its KBE (see [1, p. 3]).
First we need some definitions and the starting point of this problem.
Consider a stochastic process of $m$ particles moving on a line. Denote their initial positions by $x \in \mathbb{R}^{m}$. The chance that the particles are at the positions $y \in \mathbb{R}^{m}$ at time $t,(0 \leq t<\infty)$, is given by the transition probability density (TPD) $p(t, y \mid x)$. We denote the trajectory of the particles by $x(t), 0 \leq t \leq \infty$. Then

$$
P[x(t)=y \mid x(0)=x]=p(t, y \mid x)
$$

Since $p(t, y \mid x)$ is a probability density we must have that

$$
\int_{\mathbb{R}^{m}} p(t, y \mid x) d y=1, \quad \forall t, x
$$

A process is a Markov process if

$$
P\left[x\left(t_{2}\right)=y_{2} \mid x\left(t_{1}\right)=y_{1}, x\left(t_{0}\right)=y_{0}\right]=P\left[x\left(t_{2}\right)=y_{2} \mid x\left(t_{1}\right)=y_{1}\right]
$$

for $t_{0}<t_{1}<t_{2}$ and $y_{0}, y_{1}, y_{2} \in \mathbb{R}^{m}$, so the probability is independent of older states, but only depends on the most recent state. Each process described by a TPD is a Markov procress because

$$
\begin{aligned}
P\left[x\left(t_{2}\right)=y_{2} \mid x\left(t_{1}\right)=y_{1}, x\left(t_{0}\right)=y_{0}\right] & =\frac{P\left[x\left(t_{2}\right)=y_{2}, x\left(t_{1}\right)=y_{1}, x\left(t_{0}\right)=y_{0}\right]}{P\left[x\left(t_{1}\right)=y_{1}, x\left(t_{0}\right)=y_{0}\right]} \\
& =\frac{p\left(t_{2}-t_{1}, y_{2} \mid y_{1}\right) p\left(t_{1}, y_{1} \mid y_{0}\right)}{p\left(t_{1}, y_{1} \mid y_{0}\right)} \\
& =p\left(t_{2}-t_{1}, y_{2} \mid y_{1}\right) \\
& =P\left[x\left(t_{2}\right)=y_{2} \mid x\left(t_{1}\right)=y_{1}\right]
\end{aligned}
$$

Next we can consider some differential equation on $\mathbb{R}^{m} \times \mathbb{R}^{+}$given by

$$
\frac{\partial}{\partial t} f(x, t)-q\left(\frac{\partial}{\partial x}\right) f(x, t)=0
$$

for $q \in P\left(\mathbb{R}^{m}\right)$. If we can find a solution $f(x, t)$ that is normalized such that $\int_{\mathbb{R}^{m}} f(x, t) d x=1$, then this solution is the TPD of the Markov process with this differential equation as its KBE.
The m-dimensional Brownian motion has the heat equation as its Kolmogorov backward equation [1, p. 3]. The Green function of the Heat equation is also the TPD of the mdimensional Brownian motion, with the initial condition $p(x \mid y, 0)=\delta(x-y)$.
The normalized solution for the Dunkl heat equation is given by $\int_{\mathbb{R}^{m}} \Gamma(x, y, t) h^{2}(y) d y=1$, it has the required normalization and so $p_{k}(t, x \mid y)=\Gamma_{k}(x, y, t)|h(y)|^{2}$ is the TPD of the Markov process with the Dunkl heat equation as its KBE. We call the associated process a Dunkl process of type $R$, with parameter $k(\alpha)$. Again we are in particular interested in the Dunkl process of type $A_{m-1}$, with parameter $k$.
Finally we consider Dyson's model of brownian motion, which is the Markov process with

$$
\begin{equation*}
\frac{\partial}{\partial t}-\sum_{i=1}^{m} \frac{\partial^{2}}{\partial x_{i}^{2}}+\frac{\beta}{2} \sum_{i=1}^{m} \sum_{j=1, j \neq i}^{m} \frac{1}{x_{i}-x_{j}} \frac{\partial}{\partial x_{i}}, \tag{12.1}
\end{equation*}
$$

as KBE.
Next we look at the Dunkl process with a symmetric initial condition. By using the symmetric distribution $\mu_{x}^{z}=\sum_{w \in S_{n}} \delta(z-w x)$, we write

$$
p_{k}^{s}(t, y \mid x)=\int_{\mathbb{R}^{m}} \sum_{w \in S_{m}} p_{k}^{s}(t, y \mid x) \mu_{z}^{x} d z=\sum_{w \in S_{m}} p_{k}^{s}(t, y \mid w x)
$$

where $p_{k}(t, y \mid x)$ is the TPD of the Dunkl process of type $A_{m}$.
For $w \in S_{m}$ we have that

$$
\begin{align*}
\frac{\partial}{\partial t} p_{k}(t, y \mid w x) & =\frac{1}{2} \sum_{i=1}^{m} \frac{\partial^{2}}{\partial x_{i}^{2}} p_{k}(t, y \mid w x)+\sum_{i=1}^{n} \sum_{j=1 ; j \neq i}^{n} \frac{k}{x_{i}-x_{j}} \frac{\partial}{\partial x_{i}} p_{k}(t, y \mid w x) \\
& -\frac{k}{2} \sum_{i=1}^{n} \sum_{j=1 ; j \neq i}^{n} \frac{p_{k}(t, y \mid w x)-p_{k}\left(t, y \mid \sigma_{i j} w x\right)}{\left(x_{i}-x_{j}\right)^{2}} \tag{12.2}
\end{align*}
$$

and by setting $z=w x$, we get

$$
\begin{aligned}
\frac{\partial}{\partial t} p_{k}(t, y \mid z) & =\frac{1}{2} \sum_{i=1}^{m} \frac{\partial^{2}}{\partial z_{w(i)}^{2}} p_{k}(t, y \mid z)+\sum_{i=1}^{n} \sum_{j=1 ; j \neq i}^{n} \frac{k}{z_{w(i)}-z_{w(j)}} \frac{\partial}{\partial z_{w(i)}} p_{k}(t, y \mid z) \\
& -\frac{k}{2} \sum_{i=1}^{n} \sum_{j=1 ; j \neq i}^{n} \frac{p_{k}(t, y \mid z)-p_{k}\left(t, y \mid \sigma_{w(i) w(j)} z\right)}{\left(z_{w(i)}-z_{w(j)}\right)^{2}}, \\
\frac{\partial}{\partial t} p_{k}(t, y \mid z) & =\frac{1}{2} \sum_{i=1}^{m} \frac{\partial^{2}}{\partial z_{i}^{2}} p_{k}(t, y \mid z)+\sum_{i=1}^{n} \sum_{j=1 ; j \neq i}^{n} \frac{k}{z_{i}-z_{j}} \frac{\partial}{\partial z_{i}} p_{k}(t, y \mid z) \\
& -\frac{k}{2} \sum_{i=1}^{n} \sum_{j=1 ; j \neq i}^{n} \frac{p_{k}(t, y \mid z)-p_{k}\left(t, y \mid \sigma_{i) j} z\right)}{\left(z_{i}-z_{j}\right)^{2}}
\end{aligned}
$$

where we have changed the summation indices to $w(i)$ and $w(j)$, which only rearranges the terms in the sum. This shows that $p_{k}(t, y \mid w x)$ is another way to write the TPD of the Dunkl heat equation and $p_{k}^{s}(t, y \mid x)=m!p_{k}(t, y \mid x)$.
Comparing equation (12.1) with the definition of $\Delta_{k}$ and the Dunkl heat equation of type $A_{n}$, we can see that the TPD of a symmetric Dunkl process of parameter k , solves the Kolmogorov backward equation of Dysons model with parameter $\beta=2 k$, since the $f(x)-f\left(\sigma_{i j} x\right)=0, \forall i, j$ if f is symmetric. For two ordered vectors, such that $x_{i}<x_{j}$ and $y_{i}<y_{j}$ for $i<j$ the TPD of Dysons model with parameter $\beta$, is given by

$$
\begin{equation*}
P_{\beta}(t, y \mid x)=\frac{m!e^{-\left(x^{2}+y^{2}\right) / 2 t}}{(2 \pi t)^{m / 2}} \prod_{j=1}^{m}\left[\frac{\Gamma(1+\beta / 2)}{\Gamma(1+j \beta / 2)}\right]\left|\theta_{m}\left(\frac{y}{\sqrt{t}}\right)\right|_{0}^{\beta} \mathcal{F}_{0}^{(2 / \beta)}\left(\frac{x}{\sqrt{t}}, \frac{y}{\sqrt{t}}\right), \tag{12.3}
\end{equation*}
$$

where $\Gamma(x)$ is the gamma function, $\theta_{m}(y)=\prod_{1 \leq i<j \leq m}\left(y_{i}-y_{j}\right)$ is the Vandermonde-determinant and ${ }_{0} \mathcal{F}_{0}^{(2 / \beta)}$ is the generalized hypergeometric function. (see [1, p.5].)

Before we can continue with the definition of the Jack polynomials and the generalized hypergeometric function, we need some theory about partitions and symmetric polynomials. For this we will use results from Chapters 4 and $B$ of [1] and [7].

Definition 12.1. [7] A permutation $\tau$ is an integer valued vector $\left(\tau_{1}, \tau_{2} \ldots, \tau_{s}\right), \tau_{i} \in \mathbb{Z}_{+}$, such that $\tau_{i} \geq \tau_{i+1}$. Define $|\tau|=\sum_{i=1}^{s} \tau_{i}, \tau!=\prod_{i=1}^{m} \tau_{i}$ ! and $l(\tau)=s$. We also use the notation $\tau \dashv n$ for $|\tau|=n$.

For a partition $\tau$, we say that $(i, j) \in \tau$ if $\tau_{i} \geq j$, for $i, j \in \mathbb{Z}^{+}$. We define the conjugate partition $\tau^{*}$ as the partition such that $(j, i) \in \tau^{*}$ if and only if $(i, j) \in \tau$. Finally we define two constants given by

$$
\eta_{\tau}=\prod_{(i, j) \in \lambda}\left((1 / k)\left(\lambda_{i}-j\right)+\lambda_{j}^{*}-i+1\right)
$$

and

$$
\eta_{\tau}^{\prime}=\prod_{(i, j) \in \lambda}\left((1 / k)\left(\lambda_{i}-j+1\right)+\lambda_{j}^{*}-i\right) .
$$

We can use the multi-index notation with respect to $\tau, l(\tau) \leq m$ which gives $x^{\tau}=\prod_{i=1}^{l(\tau)} x_{i}^{\tau_{i}}$. If $l(\tau)<m$, we get a vector $\tau^{\prime}$ of length $m$ by writing $\tau_{1}=\tau_{1}^{\prime}, \ldots, \tau_{l(\tau)}=\tau_{l(\tau)}^{\prime}$ and $\tau_{l(\tau)+1}^{\prime}=$ $\cdots=\tau_{m}^{\prime}=0$. Note that $x^{\tau}=x^{\tau^{\prime}}$. For a permutation $\sigma \in S_{m}$ define $\sigma(\tau)=\left(\tau_{\sigma(1)}^{\prime}, \ldots, \tau_{\sigma(m)}^{\prime}\right.$. Define by $M(\tau, m)$ the number of distincts permutations of $\tau^{\prime}$. To find this number, look at $\tau^{\prime}$ as $m$-dimensional vector. Assume $\tau^{\prime}$ has $\rho$ distinct values including 0 , with multiplicity $l_{i}^{\tau}, 1 \leq i \leq \rho$, then

$$
\begin{equation*}
M(\tau, m)=\frac{m!}{l_{1}^{\tau}!\ldots l_{\rho}^{\tau}!} . \tag{12.4}
\end{equation*}
$$

Definition 12.2. [7, Rem. 2.8] Let $\tau$ be a partition with $l(\tau) \leq m$. The monomial symmetric function is defined by

$$
m_{\tau}=\sum_{\substack{\sigma \in S_{m} \\ \sigma(\tau) \text { distinct }}} x^{\sigma\left(\tau^{\prime}\right)},
$$

and is also given by

$$
m_{\tau}=\sum_{\sigma \in S_{m}} \frac{x^{\sigma\left(\tau^{\prime}\right)}}{l_{1}^{\tau}!\ldots l_{\rho}^{\tau}!} .
$$

We can write

$$
\exp \left(x_{1}+\cdots+x_{m}\right)=\sum_{n=0}^{\infty} \frac{1}{n!}\left(x_{1}+\cdots+x_{m}\right)^{n}=\sum_{n=0}^{\infty} \frac{1}{n!} \sum_{\substack{\tau \dashv n \\ l(\tau) \leq m}} \frac{n!}{\tau!l_{1}^{\tau}!\ldots l_{\rho}^{\tau!}} m_{\tau} .
$$

Definition 12.3. [7, Def. 2.5] For two partitions $\mu, \nu$ with the same norm and $t=\max (l(\mu), l(\nu)$, we say that $\mu \preceq \nu$ if

$$
\sum_{i=1}^{j} \mu_{i} \leq \sum_{i=1}^{j} \nu_{i}, \forall j<t
$$

and

$$
\sum_{i=1}^{t} \mu_{i}=\sum_{i=1}^{t} \nu_{i} .
$$

If any of the inequalities is strict, we say that $\mu \prec \nu$. Note that this only a partial ordering.
For any real-valued matrix $A_{\mu, \nu}$, such that $A_{\mu, \nu} \neq 0$ if and only if $\mu \preceq \nu$, we can define a set of symmetric polynomials defined by $m_{\tau, A}=\sum_{\tau} A_{\tau, \nu} m_{\nu}$. The Jack polynomials are a special type of these symmetric polynomials. (See [7, Def 2.9].)
Definition 12.4. [7, Eqn. (4)] The generalized shifted factorial, for a parameter $\alpha \in \mathbb{R}$ and a partition $\tau$, is denoted by

$$
(a)_{\tau}^{\alpha}=\prod_{i=1}^{l(\tau)} \frac{\Gamma\left(a-(i-1) / \alpha+\tau_{i}\right)}{\Gamma(a-(i-1) / \alpha)}
$$

Definition 12.5. [7, Def 2.10] The $\mathcal{C}$-normalized Jack polynomial $\mathcal{C}_{\tau}^{\alpha}$ is defined as the only polynomial homogeneous eigenfunction of the operator

$$
D^{*}=\sum_{i=1}^{m} x_{i}^{2} \frac{d^{2}}{d x_{i}^{2}}+\frac{2}{\alpha} \sum_{1 \leq i \neq j \leq m} \frac{x_{i}^{2}}{x_{i}-x_{j}} \frac{d}{d x_{i}},
$$

with eigenvalue $\sum_{i=1}^{m} \tau_{i}\left(\tau_{i}-1-\frac{2}{\alpha}(i-1)\right)+n(m-1)$ having leading term corresponding to $m_{\tau}$. In addition the functions are normalized by

$$
\sum_{\tau \vdash n, l(\tau) \leq m} \mathcal{C}_{\tau}^{\alpha}\left(x_{1}, \ldots, x_{m}\right)=\left(x_{1}+\cdots+x_{m}\right)^{l}
$$

Definition 12.6. [1, Eqn. B.9] The $\mathcal{P}$-normalized Jack polynomials $\mathcal{P}_{\lambda}^{\alpha}$ are the Jack functions which are normalized such that coefficient in front of the leading term is 1 . They are defined by

$$
\mathcal{P}_{\tau}^{\alpha}=\left(\prod_{(i, j) \in \tau} \alpha\left(\tau_{i}-1+j\right)+\tau_{j}^{*}-i\right) \frac{1}{\alpha^{|\tau|}|\tau|!} \mathcal{C}_{\tau}^{\alpha} .
$$

Definition 12.7. We can define the $u_{\tau \lambda}(\alpha)$, where $\tau$ and $\lambda$ run over the partition indices, as the matrix such that

$$
\mathcal{P}_{\tau}^{\alpha}(x)=\sum_{\substack{\lambda \preceq \tau \\|\lambda|=|\tau|}} u_{\tau \lambda}(\alpha) m_{\lambda}(x) .
$$

Definition 12.8. [7, Def 2.22] The formal definition of the generalized hypergeometric function with the parameters $a_{1}, \ldots, a_{p}$ and $b_{1}, \ldots, b_{q}$ on the variables $x_{1}, \ldots, x_{m}$ is given by

$$
{ }_{p} \mathcal{F}_{q}^{\alpha}\left(a_{1}, \ldots, a_{p} ; b_{1}, \ldots, b_{q} ; x_{1}, \ldots, x_{m}\right)=\sum_{n=0}^{\infty} \sum_{\tau \dashv n} \frac{\left(a_{1}\right)_{\tau} \ldots\left(a_{p}\right)_{\tau}}{n!\left(b_{1}\right)_{\tau} \ldots\left(b_{q}\right)_{\tau}} C_{\tau}^{\alpha}\left(x_{1}, \ldots, x_{m}\right) .
$$

This definition of a hypergeometric function assumes an argument $\left(x_{1}, \ldots x_{m}\right) \in \mathbb{R}^{m}$. We can extend the definition to hypergeometric functions of arguments in $\left(x_{1}, \ldots, x_{m} ; y_{1}, \ldots, y_{m} ; \ldots\right) \in$ $\mathbb{R}^{m} \times \mathbb{R}^{m} \times \ldots$ by inserting an additional $C_{\lambda}^{\alpha}\left(y_{1}, \ldots y_{m}\right) / C_{\lambda}^{\alpha}(1, \ldots, 1)$ for each extra vector in $\mathbb{R}^{m}$. This gives

$$
\begin{equation*}
{ }_{0} \mathcal{F}_{0}^{(1 / k)}(x, y)=\sum_{n=0}^{\infty} \sum_{\lambda \dashv n} \frac{\mathcal{C}_{\lambda}^{(1 / k)}(x) \mathcal{C}_{\lambda}^{(1 / k)}(y)}{k!\mathcal{C}_{\lambda}^{(1 / k)}(1)} \tag{12.5}
\end{equation*}
$$

We want to rewrite (12.5) in terms of the $\mathcal{P}$-normalized Jack functions. This gives ([1] Eqn. B.10]

$$
\begin{equation*}
{ }_{0} \mathcal{F}_{0}^{(1 / k)}(x, y)=\sum_{n=0}^{\infty} \sum_{\lambda \dashv n} \frac{\eta_{\tau} \mathcal{P}_{\lambda}^{(1 / k)}(x) \mathcal{P}_{\lambda}^{(1 / k)}(y)}{\eta_{\tau}^{\prime}(k m)_{\tau}^{(1 / k)}} . \tag{12.6}
\end{equation*}
$$

Theorem 12.9. [1, Thm. 2] The effect of the intertwining operator $V_{k}$ of type $A_{m-1}$ on a monomial symmetric function $m_{\lambda}(x)$ in $m$ variables is given by

$$
\sum_{\lambda} \frac{\left(u^{-1}\right)_{\lambda \tau}(1 / k)}{\lambda!M(\lambda, m)} V_{k} m_{\lambda}(x)=\frac{\eta_{\tau}(1 / k) \mathcal{P}_{\tau}^{(1 / k)}(x)}{\eta_{\tau}^{\prime}(1 / k)(k m)_{\tau}^{1 / k)}} .
$$

Proof. By rescaling equation (10.5) and combining this with equations (12.3) and 12.6) we find

$$
\begin{align*}
\sum_{w \in s_{n}} K(w x, y) & =m!{ }_{0} \mathcal{F}_{0}^{1 / k}(x, y) \\
& =m!\sum_{n=0}^{\infty} \sum_{\lambda \dashv n} \frac{\eta_{\tau} \mathcal{P}_{\lambda}^{(1 / k)}(x) \mathcal{P}_{\lambda}^{(1 / k)}(y)}{\eta_{\tau}^{\prime}(k m)_{\tau}^{(1 / k)}} \tag{12.7}
\end{align*}
$$

We continue by expanding the symmetric exponential into the monomial symmetric functions which gives

$$
\begin{aligned}
\sum_{w \in S_{m}} \exp (\langle w x, y\rangle) & =\sum_{w \in S_{m}} \sum_{n=0}^{\infty} \sum_{\substack{\lambda \rightarrow n}} \frac{1}{\lambda!} \sum_{\tau \in S_{m}} \prod_{j=1}^{m}\left(x_{w(j)} y_{j}\right)^{\lambda_{\tau(j)}}, \\
& =\sum_{l(\lambda) \leq m} \frac{1}{\lambda!} \sum_{\tau \in S_{m}}\left\{\sum_{w \in S_{m}} \prod_{j=1}^{m} x_{w(j)}^{\lambda_{\tau(j)}}\right\} \prod_{j=1}^{m} y_{j}^{\lambda_{\tau(j)}}, \\
& =\sum_{l(\lambda) \leq m} \frac{1}{\lambda!}\left\{\sum_{w^{\prime} \in S_{m}} \prod_{j^{\prime}=1}^{m} x_{j^{\prime}}^{\lambda_{w^{\prime}\left(j^{\prime}\right)}}\right\} \sum_{\tau \in S_{m}} \prod_{j=1}^{m} y_{j}^{\lambda_{\tau(j)}} .
\end{aligned}
$$

By definition the term on the right is equal to $m_{\lambda}(y)$ and the term inside braces is equal to $m_{\lambda}(x)$ multiplied by the number of non-distinct permutations of $\lambda$. Using equation (12.4), we can write

$$
\sum_{w \in S_{m}} \exp (\langle w x, y\rangle)=\sum_{\lambda} \frac{m!m_{\lambda}(x) m_{\lambda}(y)}{\lambda!M(\lambda, m)} .
$$

By inserting the inverse of Definition 12.7 after applying $V_{k}$, we get

$$
\sum_{w \in S_{m}} K(w x, y)=V_{k}^{x} \sum_{\lambda} \frac{m!m_{\lambda}(x)}{\lambda!M(\lambda, m)} \sum_{\nu}\left(u^{-1}\right)_{\lambda \nu}(1 / k) \mathcal{P}_{\nu}^{(1 / k)}(y) .
$$

By using this, equation 12.6 and equation 12.7 we find

$$
V_{k}^{x} \sum_{\lambda} \frac{m!m_{\lambda}(x)}{\lambda!M(\lambda, m)} \sum_{\nu}\left(u^{-1}\right)_{\lambda \nu}(1 / k) \mathcal{P}_{\nu}^{(1 / k)}(y)=\sum_{\tau} \frac{\eta_{\tau}(1 / k) \mathcal{P}_{\tau}^{(1 / k)}(x) \mathcal{P}_{\tau}^{(1 / k)}(y)}{\eta_{\tau}^{\prime}(1 / k)(k m)_{\tau}^{(1 / k)}},
$$

and by using the orthogonality relations of the Jack polynomials and the linearity of $V_{k}$, we equate the coefficients of the same Jack polynomials in $y$, which gives

$$
\sum_{\lambda} \frac{\left(u^{-1}\right)_{\lambda \tau}(1 / k)}{\lambda!M(\lambda, m)} V_{k} m_{\lambda}(x)=\frac{\eta_{\tau}(1 / k) \mathcal{P}_{\tau}^{(1 / k)}(x)}{\eta_{\tau}^{\prime}(1 / k)(k m)_{\tau}^{(1 / k)}},
$$

which proves the theorem.

## Chapter 13

## Arbitrary linear operators of degree -1

In the previous chapters we have found that we can generalize almost every aspect of harmonic analysis on polynomials to Dunkl harmonic analysis. Basically this is done by applying the intertwining operator at the appropriate place, using the generalized measures $|h|^{2} d \mu$ instead of $d \mu$ and the Dunkl dimension $(m+2 \gamma)$ instead of $m$.
In this chapter we will investigate which results can be generalized to arbitrary operators of degree $\pm 1$ and we will eventually see, that Dunkl operators are kind of unique.

Let $V, W$ be finite dimensional linear spaces. Let $A: V \rightarrow W$ and $B: W \rightarrow V$ be linear maps. We want to find conditions on $A, B$ such that there exists an inner product $\langle\cdot, \cdot\rangle_{V}$ on $V$, an inner product $\langle\cdot, \cdot \cdot\rangle_{W}$ on $W$ and $\langle A v, w\rangle_{W}=\langle v, B w\rangle_{V}$, for all $v \in V, w \in W$. For this we need a few lemmas.
First recall Lemma 4.6 on page 17 , which is restated in the next lemma.
Lemma 13.1. Let $V, W$ be finite dimensional linear spaces with positive definite inner product, $A: V \rightarrow W$ and $B: W \rightarrow V$ linear maps and let $A$ be the adjoint of $B$. Then we have that $V=\operatorname{im}(B) \oplus \operatorname{ker}(A)$. We also have that $W=\operatorname{im}(A) \oplus \operatorname{ker}(B)$.

This lemma was already proven on page 17 .
Let $V$ and $W$ be finite dimensional linear spaces. Let $A: V \rightarrow W$ and $B: W \rightarrow V$ be linear operators. Let $V$ have an inner product $\langle\cdot, \cdot\rangle_{V}$. We want to find conditions on $A$ and $B$, such that there is an inner product $\langle\cdot, \cdot\rangle_{W}$ on $W$, with $\langle A x, y\rangle=\langle x, B y\rangle$, for all $x \in V, y \in W$, so $A$ and $B$ are adjoints with respect to these two inner products.

Lemma 13.2. Let $V$ and $W$ be finite dimensional linear spaces. Let $A: V \rightarrow W$ be a linear operator and let $B: W \rightarrow V$ be a linear operator, such that $V=\operatorname{ker}(A) \oplus \operatorname{im}(B)$ and $W=\operatorname{ker}(B) \oplus \operatorname{im}(A)$.
Then the operator $A B$ is a bijection from $\operatorname{im}(A)$ to $\operatorname{im}(A)$
Proof. By the condition $W=\operatorname{ker}(B) \oplus \operatorname{im}(A)$, we see that $\left.B\right|_{\operatorname{im}(A)}$ is injective and it follows that $B: \operatorname{im}(A) \rightarrow \operatorname{im}(B)$ is bijective. In the same way we see that $A: \operatorname{im}(B) \rightarrow \operatorname{im}(A)$ is bijective.
By combining these two results we see that $A B: \operatorname{im}(A) \rightarrow \operatorname{im}(A)$ is a bijection.

Lemma 13.3. Let $V, W$ be finite dimensional vector spaces. Let $A: V \rightarrow W$ and $B: W \rightarrow V$ be linear maps. Then the following two statements are equivalent:
(i) There exists an inner product $\langle\cdot, \cdot\rangle_{V}$ on $V$ and there exists an inner product $\langle\cdot, \cdot\rangle_{W}$ on $W$, such that $\langle A v, w\rangle_{W}=\langle v, B w\rangle_{V}$, for all $v \in V, w \in W$.
(ii) The spaces $V$ and $W$ decompose as $V=\operatorname{ker}(A) \oplus \operatorname{im}(B)$ and $W=\operatorname{ker}(B) \oplus \operatorname{im}(A)$ and the map $A \circ B$ is diagonalizable with eigenvalues $\geq 0$.

Proof. First we show that (i) implies (ii).
From Lemma 13.1 it follows that statement (i) implies the decompositions $V=\operatorname{ker}(A) \oplus \operatorname{im}(B)$ and $W=\operatorname{ker}(B) \oplus \operatorname{im}(A)$. Also if $A$ and $B$ are adjoints, then $A B=A A^{*}$ which is symmetric and semi-positive definite, so it is diagonalizable and has eigenvalues $\geq 0$. Next we show that (ii) implies (i).

From Lemma 13.2 it follows that $A B: \operatorname{im}(A) \rightarrow \operatorname{im}(A)$ is a bijection, so all eigenvalues of $\left.A \circ B\right|_{\operatorname{im}(A)}$ are non-zero, hence strictly positive. The image of $A$ has dimension $k$. Let $\left\{e_{i}\right\}_{1 \leq i \leq k}$ be a basis of eigenvectors of $\left.A B\right|_{\operatorname{im}(A)}$. Let $\langle\cdot, \cdot\rangle_{\mathrm{im}(A)}$ be the inner product such that $\left\{e_{i}\right\}$ is an orthonormal basis, so

$$
\left\langle e_{i}, e_{j}\right\rangle_{\operatorname{im}(A)}=\delta_{i j}, \text { for } 1 \leq i, j \leq k
$$

We have the relation $A \circ B e_{i}=\lambda_{i} e_{i}$. By multiplying both sides with $B$, this gives $B \circ A \circ B e_{i}=$ $\lambda_{i} B e_{i}$. Denote $f_{i}=B e_{i}$. The elements $\left\{f_{i}\right\}_{1 \leq i \leq k}$ form a basis of $\operatorname{im}(B)$.
Let $\langle\cdot, \cdot\rangle_{\mathrm{im}(B)}$ be the inner product such that $\left\langle f_{i}, f_{j}\right\rangle_{\operatorname{im}(B)}=\mu_{i} \delta_{i j}$, for $\mu_{i} \in \mathbb{R}$. We want to choose the constants $\mu_{i}$ such that $\left\langle A e_{i}, f_{j}\right\rangle_{\mathrm{im}(B)}=\left\langle e_{i}, B f_{j}\right\rangle_{\mathrm{im}(A)}$. For this note that

$$
\left\langle A e_{i}, f_{j}\right\rangle_{\operatorname{im}(B)}=\left\langle f_{i}, f_{j}\right\rangle_{\operatorname{im}(B)}=\mu_{i} \delta_{i j}
$$

and

$$
\left\langle e_{i}, B f_{i}\right\rangle_{\operatorname{im}(A)}=\left\langle e_{i}, B A e_{j}\right\rangle_{\operatorname{im}(A)}=\lambda_{i} \delta_{i j},
$$

so it follows that $\mu_{i}=\lambda_{i}$.
Next choose an inner product $\langle\cdot, \cdot\rangle_{\operatorname{ker}(A)}$ on $\operatorname{ker}(A)$ and choose an inner product $\langle\cdot, \cdot\rangle_{\operatorname{ker}(B)}$ on $\operatorname{ker}(B)$. Define the inner product $\langle\cdot, \cdot\rangle_{V}$ by $\langle\cdot, \cdot\rangle_{\operatorname{ker}(A)} \oplus\langle\cdot, \cdot\rangle_{\mathrm{im}(B)}$ and the inner product $\langle\cdot, \cdot\rangle_{W}$ by $\langle\cdot, \cdot\rangle_{\operatorname{ker}(A)} \oplus\langle\cdot, \cdot\rangle_{\operatorname{im}(B)}$. Then $A$ and $B$ are adjoints with respect to these inner products.

Definition 13.4. [17, Thm. 2.53] We denote by $C^{\infty}\left(\mathbb{R}^{m}\right) \otimes \Lambda\left(\mathbb{R}^{m}\right)$ the graded algebra of smooth complex valued forms on $\mathbb{R}^{m}$.
The exterior derivative $d: C^{\infty}\left(\mathbb{R}^{m}\right) \otimes \Lambda\left(\mathbb{R}^{m}\right) \rightarrow C^{\infty}\left(\mathbb{R}^{m}\right) \otimes \Lambda\left(\mathbb{R}^{m}\right)$ is defined to be the unique operator satisfying

1. linearity, $d(a \alpha+b \beta)=a d \alpha+b d \beta$, for $a, b \in \mathbb{R}, \alpha, \beta \in C^{\infty}\left(\mathbb{R}^{m}\right) \otimes \Lambda\left(\mathbb{R}^{m}\right)$
2. for a 0 -form $f \in C^{\infty}\left(\mathbb{R}^{m}\right) \otimes \Lambda^{0}\left(\mathbb{R}^{m}\right) \simeq C^{\infty}\left(\mathbb{R}^{m}\right)$, $d f$ is the usual differential.
3. for an $\mathbf{j}$-form $\alpha$ and a l-form $\beta, d(\alpha \wedge \beta)=d(\alpha) \wedge \beta+(-1)^{j} \alpha \wedge d \beta$
4. $d(d \alpha)=0$ for all forms.

From this definition it follows that $d$ is of degree 1 . In particular we find for 0 -forms that $d: C^{\infty}\left(\mathbb{R}^{m}\right) \rightarrow C^{\infty}\left(\mathbb{R}^{m}\right) \otimes \Lambda^{1}\left(\mathbb{R}^{m}\right)$ is given by

$$
d f=\sum_{j=1}^{m} \partial_{i} f e^{i}
$$

and for 1-forms that $d: C^{\infty}\left(\mathbb{R}^{m}\right) \otimes \Lambda^{1}\left(\mathbb{R}^{m}\right) \rightarrow C^{\infty}\left(\mathbb{R}^{m}\right) \otimes \Lambda^{2}\left(\mathbb{R}^{m}\right)$ is given by

$$
d\left(\sum_{j=1}^{m} f_{j} e^{j}\right)=\sum_{j=1}^{m} \sum_{i<j}\left(\partial_{i} f_{j}-\partial_{j} f_{i}\right) e^{i} \wedge e^{j}
$$

Definition 13.5. Let $P$ be the space of polynomials on $\mathbb{R}^{m}$ and let $P_{n}$ be the space of homogeneous polynomials of degree $n$. Let $A_{i}$ be a set of commuting operators of degree -1 . Define, analogous to Definition 13.4 , the $A$-exterior derivative $d_{n, 0}^{A}: P_{n} \rightarrow P_{n-1} \otimes \Lambda^{1}\left(\mathbb{R}^{m}\right)$ by

$$
d_{n, 0}^{A}(p)=\sum_{j=1}^{m} A_{j} e^{j}
$$

and the $A$-exterior derivative $d_{n-1,1}^{A}: P_{n-1} \otimes \Lambda^{1}\left(\mathbb{R}^{m}\right) \rightarrow P_{n-2} \otimes \Lambda^{2}\left(\mathbb{R}^{m}\right)$ by

$$
d_{n-1,1}^{A}\left(\sum_{j=1}^{m} p_{j} e^{j}\right)=\sum_{j=1}^{m} \sum_{i<j}\left(A_{i} p_{j}-A_{j} p_{i}\right) e^{i} \wedge e^{j} .
$$

In particular, it follows that $d_{n-1,1} d_{n, 0} p=\sum_{j=1}^{m} \sum_{i<j}\left[\left(A_{i} A_{j}-A_{j} A_{i}\right) p\right] e^{i} \wedge e^{j}=0$, since the operators $A_{i}$ are commuting.

Lemma 13.6. Let $P$ be the space of polynomials on $\mathbb{R}^{m}$ and let $P_{n}$ be the subspace of homogeneous polynomials of degree $n$. Let $A_{i}: P \rightarrow P, 1 \leq i \leq m$ be a set of commuting operators of degree -1 , such that $\cap \operatorname{ker}\left(A_{i}\right)=P_{0}$. Then there exists a unique linear operator $V: P \rightarrow P$, such that
(i) $V(1)=1$,
(ii) $V\left(P_{n}\right) \subset P_{n}$ for all $n$,
(iii) $\partial_{i} \circ V=V \circ A_{i}$ for all $i$.

The operator $V$ is invertible.
Proof. By induction we will define linear maps $V_{n}: P_{n} \rightarrow P_{n}$, such that $V_{0}=I_{P_{0}}$ and $\partial_{i} \circ V_{n}=V_{n-1} \circ A_{i}$ on $P_{n}$.
Assume that $V_{0}, \ldots, V_{n-1}$ have already been defined. These operators induce the maps

$$
V_{n-1} \otimes I: P_{n-1} \otimes \Lambda^{1}\left(\mathbb{R}^{m}\right) \circlearrowleft
$$

and

$$
V_{n-2} \otimes I: P_{n-2} \otimes \Lambda^{2}\left(\mathbb{R}^{m}\right) \circlearrowleft
$$

and the following diagram commutes:

$$
\begin{array}{rrr}
P_{n-1} \otimes \Lambda^{1}\left(\mathbb{R}^{m}\right) & \xrightarrow{d_{n-1,1}^{A}} & P_{n-2} \otimes \Lambda^{2}\left(\mathbb{R}^{m}\right) \\
\downarrow V_{n-1} \otimes I & & \downarrow V_{n-2} \otimes I \\
P_{n-1} \otimes \Lambda^{1}\left(\mathbb{R}^{m}\right) & \xrightarrow{d_{n-1,1}} & P_{n-2} \otimes \Lambda^{2}\left(\mathbb{R}^{m}\right)
\end{array}
$$

We want of find an operator $V_{n}$ such that the following diagram commutes:

$$
\begin{array}{lrrrr}
P_{n} & \xrightarrow{d_{n, 0}^{A}} & P_{n-1} \otimes \Lambda^{1}\left(\mathbb{R}^{m}\right) & \xrightarrow{d_{n-1,1}^{A}} & P_{n-2} \otimes \Lambda^{2}\left(\mathbb{R}^{m}\right) \\
\vdots V_{n} & & \downarrow V_{n-1} \otimes I & \circlearrowleft & \downarrow V_{n-2} \otimes I \\
\downarrow & & \\
P_{n} & \xrightarrow{d_{n, 0}} & P_{n-1} \otimes \Lambda^{1}\left(\mathbb{R}^{m}\right) & \xrightarrow{d_{n-1,1}} & P_{n-2} \otimes \Lambda^{2}\left(\mathbb{R}^{m}\right)
\end{array}
$$

Take $p \in P_{n}$. Then $d_{n, 0}^{A}(p) \in \operatorname{ker}\left(d_{n-1,1}^{A}\right)$, so

$$
d\left(\left(V_{n-1} \otimes I\right) d^{A}(p)\right)=\left(V_{n-2} \otimes I\right)\left(d^{A} d^{A} p\right)=0
$$

Since each closed $k$-form on $\mathbb{R}^{m}$ is exact, we have that $\operatorname{ker}\left(d_{n-2,1}\right)=\operatorname{im}\left(d_{n-1,0}\right)$. This means there exists a $q \in P_{n}$, such that $d_{n, 0}(q)=\left(V_{n-1} \otimes I\right) d_{n, 0}^{A}(p)$. This $q$ is unique modulo $\operatorname{ker}\left(d_{n, 0}\right)=P_{n} \cap P_{0}=0$, so $q$ is unique. This means we can define the map $V_{n}: P_{n} \rightarrow P_{n}$ by $V_{n}(p)=q . V_{n}$ is clearly linear. We also have that $d_{n, 0}\left(V_{n}(p)\right)=d_{n, 0}(q)=\left(V_{n-1} \otimes I\right) d_{n, 0}^{A}(p)$, which shows that $\partial_{j} \circ V_{n}=V_{n-1} \circ A_{j}$ on $P_{n}$.
Next suppose that there are $p$ and $p^{\prime}$, such that $V_{n}(p)=V_{n}\left(p^{\prime}\right)=q$.
Then $d_{n, 0}(q)=\left(V_{n-1} \otimes I\right) d_{n, 0}^{A}(p)$ and $d_{n, 0}(q)=\left(V_{n-1} \otimes I\right) d_{n, 0}^{A}\left(p^{\prime}\right)$, so $\left.V_{n-1} \otimes I\right) d_{n, 0}^{A}\left(p-p^{\prime}\right)=0$ so $p-p^{\prime} \in\left(\operatorname{ker}\left(d^{A}\right) \cap P_{n}\right)=\left(\cap_{j=1}^{m}\left(\operatorname{ker} A_{i}\right)\right) \cap P_{n}=0$, so $p=p^{\prime}$ and $V_{n}$ is injective. Since $V_{n} \in \operatorname{End}\left(P_{n}\right)$ this means that $V_{n}$ is invertible.
So by induction the maps $V_{n}: P_{n} \rightarrow P_{n}$, such that $V_{0}=I_{P_{0}}$ and $\partial_{i} \circ V_{n}=V_{n-1} \circ A_{i}$ on $P_{n}$. Furthermore the maps $V_{n}$ are invertible and unique.
Define $V: P \rightarrow P$, by $V(p)=V\left(\sum_{i=0}^{\operatorname{deg}(p)} p_{i}\right)=\sum_{i=0}^{\operatorname{deg}(p)} V_{i}\left(p_{i}\right)$, with $p_{i} \in P_{i}$. The map $V$ is invertible and unique and has properties (i)-(iii) stated in the lemma, because $\left.V\right|_{P_{n}}=V_{n}, n \in$ $\mathbb{N}$ is invertible and unique and has properties (i)-(iii) stated in the lemma.

We call the map $V$ from the previous lemma the $(\partial, A)$-intertwining operator, or in short the $A$-intertwining operator. In the following we shall denote this map by $V_{A}$ or $V_{\partial, A}$.

Corollary 13.7. Let $P$ be the space of polynomials on $\mathbb{R}^{m}$ and let $P_{n}$ be the subspace of homogeneous polynomials of degree $n$. Let $A_{i}: P \rightarrow P, 1 \leq i \leq m$ be a set of commuting operators of degree -1 , such that $\cap \operatorname{ker}\left(A_{i}\right)=P_{0}$. Let $B_{i}: P \rightarrow P, 1 \leq i \leq m$ be another set of commuting operators of degree -1 , such that $\cap \operatorname{ker}\left(B_{i}\right)=P_{0}$. Then the map $V_{A, B}=V_{A, \partial} V_{B, \partial}^{-1}$ is the unique $A, B$-intertwining operator, which means that
(i) $V_{A, B}(1)=1$,
(ii) $V_{A, B}\left(P_{n}\right) \subset P_{n}$ for all $n$,
(iii) $A_{i} \circ V_{A, B}=V_{A, B} \circ B_{i}$ for all i.

Its inverse is given by $V_{B, A}=V_{B, \partial} V_{A, \partial}^{-1}$.

Proof. By Lemma 13.6 the maps $V_{A, \partial}$ and $V_{B, \partial}$ exists and are invertible. Properties (i) and (ii) and follow in a trivial way. The operator $V_{A, B}$ is invertible because $V_{A, \partial}$ and $V_{B, \partial}^{-1}$ are invertible.
For (iii) note that

$$
A_{i} \circ V_{A, B}=V_{A, \partial} \circ \partial_{i} \circ V_{B, \partial}^{-1}=V_{A, B} \circ B_{i} .
$$

Let $R \subset \mathbb{R}^{m}$ be a root system with a weight function $k$. Let $e_{i}, 1 \leq i \leq m$ be an orthonormal basis of $\mathbb{R}^{m}$. Then the Dunkl operators $T_{i}: P \rightarrow P, 1 \leq i \leq m$ are a commuting set of operators of degree -1 (see Definition 6.3 and Theorem 6.11). If $k$ is nondegenerate (Definition 7.19), there exist a unique ( $T, \partial$ )-intertwining operator $V_{T, \partial}$ on $P$, by 13.6. The operator $V_{T, \partial}$ is equal to the operator $V$ in (7.8).
A special case of Lemma 13.6, is given by $A_{i}=T_{i}$, where $V_{A}$ is the inverse of the operator $V$ given in 7.8).

Definition 13.8. Let the operators $A_{i}$ be a set of commuting operators as in Lemma 13.6 , with the associated intertwining operator $V_{A}$. We can define the set of $A$-monomials by $z_{\alpha}(x)=V_{A} x^{\alpha}$, for $x^{\alpha} \in P\left(\mathbb{R}^{m}\right)$ and $x \in \mathbb{R}^{m}$. We can also define the kernel $K_{A}(x, y), x, y \in$ $\mathbb{R}^{m}$ by

$$
K_{A}(x, y) \equiv \sum_{n=0}^{\infty} K_{A, n}(x, y)=\sum_{n=0}^{\infty} \sum_{\alpha=|n|} 1 / n!V_{A}^{x}\left(\langle x, y\rangle^{n}\right)=\sum_{n=0}^{\infty} \sum_{\alpha=|n|} \frac{z_{\alpha}(x) y^{\alpha}}{\alpha!}
$$

Although we can always work with the function $K_{A, n}$, the kernel $K_{A}$ only makes sense, if the sum converges. If this is the case, we also have that $K_{A}(x, y)=V_{A}^{x} \exp (\langle x, y\rangle)$.

Lemma 13.9. Let $p \in P_{n}$. Then $K_{A, n}\left(x, A^{y}\right) p(y)=p(x)$. This means that $K_{A, n}$ is a reproducing kernel.
Proof. Let $q \in P_{n}$, then $1 / n!\left\langle x, \partial^{y}\right\rangle^{n} q(y)=q(x)$. By applying $V_{A}^{y}$ to both sides we find

$$
\frac{\left\langle x, A^{y}\right\rangle^{n}}{n!} V_{A}^{y} q(y)=q(x),
$$

since the right hand side is constant in $y$. By applying $V_{A}^{x}$ to both sides we find

$$
K_{A, n}\left(x, A^{y}\right) V_{A}^{y} q(y)=V_{A}^{x} q(x),
$$

and by using that $V_{A}$ is one to one on $P_{n}$, we can write each $p \in P_{n}$ as $V_{A} q$, for some $q \in P_{n}$ which shows that

$$
K_{A, n}\left(x, A^{y}\right) p(y)=p(x),
$$

for all $p \in P_{n}$.
Lemma 13.10. Assume that $K_{A}(x, y)$ exists, and that $K_{A}(x, y)=K_{A}(y, x)$. Then for $p, q \in$ $P_{n}$, we have a pairing on the polynomials given by

$$
[p, q]_{A}=p\left(A^{x}\right) q(x)=K_{A, n}\left(A^{x}, A^{y}\right) p(x) q(y)
$$

We can extend this pairing to all of $P$ by defining $[p, q]_{A}=0$ if $p \in P_{n}$ and $q \in P_{m}$.
If this pairing is positive definite for all polynomials, it defines an inner product on $P\left(\mathbb{R}^{m}\right.$ and $A_{i}$ and $x_{i}$ are adjoints with respect to this pairing.

Proof. First of all, note that the pairing is linear. Next we need to show that the pairing is symmetric. Let $p, q \in P_{n}$, then

$$
p\left(A^{y}\right) q(y)=K_{A, n}\left(A^{y}, A^{x}\right) p(x) q(y)=K_{A, n}\left(A^{x}, A^{y}\right) p(x) q(y)=q\left(A^{x}\right) p(x)
$$

where we have used Lemma 13.9 and the symmetry of $K_{A, n}$.
Since positive definiteness was assumed in the lemma, the pairing defines an inner product on P.
Finally note that $\left[x^{\alpha}, A_{i} x^{\beta}\right]_{A}=x^{\alpha+e_{i}}\left(A^{x}\right) x^{\beta}=\left[x_{i} x^{\alpha}, x^{\beta}\right]_{A}$, so $x_{i}$ and $A_{i}$ are adjoints with respect to this inner product.

Example 13.11. Let $m=1$. Let $c(0)=0$ and let $c(n) \in \mathbb{R}_{+}$for $n \in \mathbb{N}_{+}$. Consider the linear function $A: P \rightarrow P$ given by $A x^{n}=c(n) x^{n-1}$.
The intertwining operator $V_{A, \partial}: P \rightarrow P$ has the properties $V_{A, \partial} 1=1$ and $V_{A, \partial} \partial=A V_{A, \partial}$. We have $V_{A, \partial} \partial^{n} x^{n}=V_{A, \partial} n!=n!$. On the other hand $A^{n} x^{n}=\prod_{i=1}^{n} c(i)$, so

$$
V_{A, \partial} x^{n}=\frac{n!}{\prod_{i=1}^{n} c(i)} x^{n} .
$$

Next let $d(0)=0$, let $d(n) \in \mathbb{R}_{+}$for $n \in \mathbb{N}_{+}$and consider the linear function $B: P \rightarrow P$ given by $B x^{n}=d(n) x^{n-1}$, then the intertwining operator $V_{A, B}$ is given by

$$
V_{A, B} x^{n}=V_{A, \partial} V_{\partial, B} x^{n}=\frac{\prod_{i=1}^{n} d(i)}{\prod_{i=1}^{n} c(i)} x^{n} .
$$

The function $A$ has associated kernels $K_{A, n}(x, y), x, y \in \mathbb{R}^{m}$ and it depends on the choice of $c$ if the sum $\sum_{n=0}^{\infty} K_{A, n}(x, y)$ converges.
Note that this example can be generalized to $m$ variables. For this we shall use the multiindex notation.
For $1 \leq i \leq m$, let $c_{i}(0)=0$ and let $c_{i}(n) \in \mathbb{R}_{+}$for $n \in \mathbb{N}_{+}$. Define the operators $A_{i}: P \rightarrow P$ by $A_{i} x^{\alpha}=c_{i}\left(\alpha_{i}\right) x^{\alpha-e_{i}}$. These operators commute because if $i \neq j$ then

$$
A_{i} A_{j} x^{\alpha}=c_{i}\left(\alpha_{i}\right) c_{j}\left(\alpha_{j}\right) x^{\alpha-e_{i}-e_{j}}=A_{j} A_{i} x^{\alpha}
$$

The intertwining operator $V_{A, \partial}$ can be found in a similar way and is given by

$$
V_{A, \partial} x^{\alpha}=\frac{\alpha!}{\prod_{i=1}^{m} \prod_{j=1}^{\alpha_{i}} c_{i}(j)} x^{\alpha} .
$$

## Recap of this chapter

In this chapter it has been shown that the existence of a intertwining operator $V: P \rightarrow P$ between partial derivatives and a set of operators $A_{i}: P \rightarrow P, 1 \leq i \leq m$ only depends on the following three properties:

- The operators $A_{i}$ are homogeneous of degree -1,
- The operators $A_{i}, 1 \leq i \leq m$ commute in $\operatorname{End}(P)$,
- The intersection of their kernels $P \cap\left(\bigcap_{i=1}^{m} \operatorname{ker}\left(A_{i}\right)\right)=P_{0}$.

The intertwining operators between the Dunkl operators and partial derivatives are examples of this type of intertwining operators.
The intertwining operators described in Lemma 13.6 might be usable to simplify calculations involving Dunkl operators and might also have applications in other mathematical fields.
As a continuation, we might try to generalize steps from Chapter 9 to construct a Fourier-like transform $\mathscr{F}_{A}: L^{2}\left(\mathbb{R}^{m}\right) \rightarrow L^{2}\left(\mathbb{R}^{m}\right)$, which has the property $\mathscr{F}_{A}\left(A_{i} f\right)(y)=y_{i} \mathscr{F}_{A}(f)$. However to be able make this generalization, we probably need to put a lot of additional constraints on the operators $A_{i}$.

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